

# Deep-Dive Studies Into Mathematics 

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# Deep Dive 

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$$
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\operatorname{Sym}^{d}(V \oplus W) & \simeq \bigoplus_{i=0}^{d} \operatorname{Sym}^{i}(V) \otimes \operatorname{Sym}^{d-i}(W) \\
\Lambda^{d}(V \oplus W) & \simeq \bigoplus_{i=0}^{d} \Lambda^{i}(V) \otimes \Lambda^{d-i}(W)
\end{aligned}
$$

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## Preface

First of all I should stress that the title page template was taken from the beautiful book [ADH21] by P. Haine et al. -the whole credit should be given to them. The following text is a compilation of notes for my studies in mathematics, that being said, it is clear that the collection of mathematical and linguistic errors is non-empty. My main references are:

- Category Theory: [Rie16] and [KS06].
- Algebra: [MK89], [Kim20], [Alu09] and [Lan93].
- Topology: [Lee11], [BBT20], [Mun00] and [Eng89].
- Differential Structures: [Zor15], [Zor16], [Rud76] and [Jos06].
- Combinatorics: [Die16].
- Machine Learning: [SB14], [Bis06].


## Part I

Category Theory

## Chapter 1

## Categories

### 1.1 Sets and Universes

In order to deal with size issues in the theory of categories, we shall work inside what are called universes.

Definition 1.1.1 (Universe). A universe $\mathcal{U}$ is defined as a set satisfying the following properties
(U1) $\varnothing \in \mathcal{U}$.
(U2) If $u \in \mathcal{U}$, then $u \subseteq \mathcal{U}$.
(U3) If $u \in \mathcal{U}$, then $\{u\} \in \mathcal{U}$.
(U4) If $u \in \mathcal{U}$, then $2^{u} \in \mathcal{U}$.
(U5) If we have an indexing set $I \in \mathcal{U}$, for which we associate a collection $\left\{u_{i} \in \mathcal{U}\right\}_{i \in I}$, then $\bigcup_{i \in I} u_{i} \in \mathcal{U}$.
(U6) $\mathbf{N} \in \mathcal{U}$.
As a consequence of such properties, a universe also satisfies
(U7) If $u \in \mathcal{U}$, then $\bigcup_{x \in u} x \in \mathcal{U}$.
(U8) Given $u, v \in \mathcal{U}$, we have $u \times v \in \mathcal{U}$.
(U9) If $v \in \mathcal{U}$, and we have $u \subseteq v$, then $u \in \mathcal{U}$.
(U10) If $I \in \mathcal{U}$ is an idexing set with an associated collection $\left\{u_{i} \in \mathcal{U}\right\}_{i \in I}$, then $\prod_{i \in I} u_{i} \in$ $\mathcal{U}$.

Axiom 1.1.2 (Grothendieck's axiom to ZF set theory). For any set $X$, there exists an universe $\mathcal{U}$ for which $X \in \mathcal{U}$.

For some commonly used terminology, we say that a set $x$ is a $\mathcal{U}$-set if $x \in \mathcal{U}$. Moreover, a set $x$ is called $\mathcal{U}$-small if it is isomorphic to some set $s \in \mathcal{U}$.

Orderings are another important topic when dealing with sets, we shall define now some of these concepts.

Definition 1.1.3. Let $I$ be a set. We define the following:

- An order on the set $I$ is a relation $\leqslant$ satisfying the following properties:
(a) The order is reflexive - that is, for all $i \in I$, we have $i \leqslant i$.
(b) The order is transitive - given any three elements $i, j, k \in I$ such that $i \leqslant j$ and $j \leqslant k$, it follows that $i \leqslant k$.
(c) The order is anti-symmetric - given any two elements $i, j \in I$, if $i \leqslant j$ and $j \leqslant i$, then $i=j$.
- An order is said to be directed, or filtrant, if $I$ is non-empty and, for every $i, j \in I$, there exists $k \in I$ for which $i \leqslant k$ and $j \leqslant k$.
- An order is said to be total if, given any $i, j \in I$, necessarily have at least one of the following relations: $i \leqslant j$ or $j \leqslant i$.
- The set $I$ is said to be inductively ordered if, for any totally ordered subset $J \subseteq I, J$ has an upper bound $u \in I$ for which $j \leqslant u$ for all $j \in J$
- If $I$ is ordered by the relation $\leqslant$, we define a strict relation $<$ as, given $i, j \in I$, we have $i \leqslant j$ if and only if $i \leqslant j$ and $i \neq j$.


### 1.2 Categories

Definition 1.2.1 (Category). A category C consists of the following data
(C1) A collection of objects. We say that $X$ is an object of $C$ by writing $X \in C$ or $X \in \operatorname{Obj}(C)^{1}$.
(C2) For every given pair of objects $X, Y \in C$ there exists a collection ${ }^{2}$ of morphisms $\operatorname{Mor}_{\mathrm{C}}(X, Y)$ with source $X$ and target $Y$. The collection of morphisms between objects of C is denoted $\operatorname{Mor}(\mathrm{C})$.
(C3) For every object $X \in C$, there exists an identity morphism $\operatorname{id}_{X} \in \operatorname{Mor}_{C}(X, X)$.
(C4) For every triple of given objects $X, Y, Z \in C$, there exists a composition map

$$
\operatorname{Mor}_{\mathrm{C}}(X, Y) \times \operatorname{Mor}_{\mathrm{C}}(Y, Z) \rightarrow \operatorname{Mor}_{\mathrm{C}}(X, Z)
$$

So that for given morphisms $f \in \operatorname{Mor}_{C}(X, Y)$ and $g \in \operatorname{Mor}_{C}(Y, Z)$ there exists a uniquely defined map $g f \in \operatorname{Mor}_{c}(X, Z)$ such that the following diagram commutes


[^0](C5) For every morphism $f: X \rightarrow Y$ we have that
$$
\operatorname{id}_{X} \subset X \xrightarrow{f} Y \longmapsto \operatorname{id}_{Y}
$$
so that $\operatorname{id}_{Y} f=f=f \operatorname{id}_{X}$.
(C6) Given objects $W, X, Y, Z \in C$, the following diagram commutes

that is, $h(g f)=(h g) f$.
Notation 1.2.2 (On arrows and diagrams). I'll adopt throughout this whole text a series of notations regarding morphisms, diagrams and so on, here I collect some of those: given a category C and objects $x, y, z \in \mathrm{C}$

- An arrow $x>y$ denotes a monomorphism - to be seen in Definition 1.3.3.
- An arrow $x \rightarrow y$ denotes an epimorphism - to be seen in Definition 1.3.4.
- An arrow $x \xrightarrow{\simeq} y$ denoted an isomorphism — to be seen in Definition 1.2.6. We say that $x$ is isomorphic to $y$, and write $x \simeq y$, if there exists an isomorphism $x \xrightarrow{\simeq} y$.
- Given arrows $f: x \rightarrow y$ and $g: y \rightarrow z$ in C, we denote the composition of $f$ with $g$ by the juxtaposition $g f: x \rightarrow z^{3}$.
- Let the following be a commutative diagram on C (that is, $f g=h$ ):


The dashed arrow $g: y \rightarrow z$ denotes that $g$ is the unique morphism between $y$ and $z$ in the category $C$ such that $f g=h$.

- The diagram in C

$$
x \xrightarrow[h]{\stackrel{g}{\longrightarrow}} y \xrightarrow{f} z
$$

denotes that $f g=f h$.
Definition 1.2.3 (Small). A category C is said to be small if it is composed of a set's worth of morphisms.

[^1]Definition 1.2.4 (Locally small). A category C is said to be locally small if for any objects $A, B \in \mathrm{C}$ there exists a set's worth of morphisms $A$ and $B$.

Corollary 1.2.5 (Unique identity). Given a category $C$ and an object $c \in C$, the identity $\mathrm{id}_{c} \in \operatorname{Mor}(\mathrm{C})$ is unique.

Proof. Let $f: c \rightarrow c$ be an identity of $c$, then $f=f \mathrm{id}_{c}=\mathrm{id}_{c}$.
Definition 1.2.6 (Isomorphism). Given a category C and objects $A, B \in \mathrm{C}$, we define a morphism $f \in \operatorname{Mor}(A, B)$ to be an isomorphism if and only if it has a both sided inverse, so that exists $f^{-1} \in \operatorname{Mor}(B, A)$ such that $f^{-1} f=\operatorname{id}_{A}$ and $f f^{-1}=\mathrm{id}_{B}$.

Proposition 1.2.7. Given an isomorphism $f$, its inverse is unique.
Proof. Suppose for instance that there are two such functions, $g, h \in \operatorname{Mor}(B, A)$, that act as an inverse for $f \in \operatorname{Mor}(A, B)$. Note that

$$
g=g \operatorname{id}_{B}=g(f h)=(g f) h=\operatorname{id}_{A} h=h
$$

Thus $g=h$ and therefore the inverse is indeed unique.
Definition 1.2.8 (Non-empty category). A category is said to be non-empty if the collection of objects is non-empty.

Definition 1.2.9 (Discrete category). A category is said to be discrete if all morphisms are the identity morphisms.

Example 1.2.10. Given a set $I$, one can view $I$ as a category whose objects are the elements of $I$ and morphisms are identities - that is, $\operatorname{Mor}_{I}(x, y)$ is the identity when $x=y$, while empty if $x \neq y$.

Definition 1.2.11 (Finite category). A category is said to be finite if the collection of all morphisms is a finite set. This category will be of important use when dealing with limits.

Definition 1.2.12 (Connected category). A category C is said to be connected if it is nonempty and, for every pair $x, y \in \mathrm{C}$, there exists a finite sequence of objects $\left(x_{0}, \ldots, x_{n}\right)$, $x_{j} \in C$ for all $0 \leqslant j \leqslant n$, such that $x_{0}=x, x_{n}=y$ and at least one of the collections of morphisms $\operatorname{Mor}_{\mathrm{C}}\left(x_{j}, x_{j+1}\right)$ or $\operatorname{Mor}_{\mathrm{C}}\left(x_{j+1}, x_{j}\right)$ is non-empty for every $0 \leqslant j \leqslant n-1$

Definition 1.2.13 (Monoid). A monoid is a set $M$ equipped with a binary operation $\otimes: M \times M \rightarrow M$ and a neutral element $e \in M$. The binary operation is associative and obeys the right and left unit laws, that is

$$
x \otimes(y \otimes z)=(x \otimes y) \otimes z \quad \text { and } \quad e \otimes x=x=x \otimes e .
$$

Example 1.2.14. A monoid $M$ defines a category with one object, denoted $B M$, such that $\operatorname{Obj}(\mathrm{BM})=\{*\}$ and $\operatorname{Mor}_{\mathrm{BM}}(*, *)=$ M. Composition of morphisms $f, g: * \rightarrow *$ is defined as $g f=g \otimes f$. The identity morphism is $\operatorname{id}_{*}=e$. Hence we have $e * f=f=f * e$ for all morphisms $f \in \operatorname{Mor}(\mathrm{BM})$.

Definition 1.2.15 (Groupoids). A groupoid is the name given to a category in which all of its morphisms are isomorphisms.

Definition 1.2.16 (Group). A group is a groupoid with one object.
Definition 1.2.17 (Automorphism). Given a category C and an object $A \in \mathrm{C}$, we define an automorphism of $A$ to be an isomorphism $A \rightarrow A$. The set consisting of all such automorphisms of this object is denoted $\operatorname{Aut}(A)$, the automorphisms have properties:
i. The composition of two automorphisms is an automorphism.
ii. Composition is associative.
iii. $\operatorname{id}_{A} \in \operatorname{Aut}(A)$.
iv. Every automorphism $f \in \operatorname{Aut}(A)$ has an inverse $f^{-1} \in \operatorname{Aut}(A)$.

With this, the structure $\operatorname{Aut}(A)$ is a group, for all choices of objects $A \in \mathrm{C}$.
Definition 1.2.18 (Subcategory). Let C be a category. We define $\mathrm{D} \subseteq \mathrm{C}$ as a subcategory of C if: $\operatorname{Obj}(\mathrm{D})$ is a restriction of $\operatorname{Obj}(\mathrm{C})$; for all $A \in \operatorname{Obj}(\mathrm{D})$ there exists $\mathrm{id}_{A} \in \operatorname{Mor}(\mathrm{D})$; for any $f \in \operatorname{Mor}(\mathrm{D})$ there exists $\operatorname{dom}(f), \operatorname{cod}(f) \in \operatorname{Obj}(\mathrm{D})$; for any composable pair of morphisms $f, g \in \operatorname{Mor}(\mathrm{D})$ there exists $f g \in \operatorname{Mor}(\mathrm{D})$.

Lemma 1.2.19. Any category $C$ contains a subcategory containing all of the objects and whose morphisms are only the isomorphisms. Such subcategory is called a maximal groupoid.

Proof. We are going to prove that the maximal groupoid, name it G is indeed a subcategory of C. Notice that $\operatorname{Obj}(\mathrm{G})=\operatorname{Obj}(\mathrm{C})$, moreover every identity is an isomorphism, then $\mathrm{id}_{*} \in \operatorname{Mor}(\mathrm{G})$. Let $f \in \operatorname{Mor}(\mathrm{G})$ then in particular we have $f \in \operatorname{Mor}(\mathrm{C})$ and hence $\operatorname{dom} f, \operatorname{codom} f \in \operatorname{Obj}(\mathrm{C})=\operatorname{Obj}(\mathrm{G})$. Consider $f \in \operatorname{Mor}_{G}(A, B)$ and $g \in$ $\operatorname{Mor}_{G}(B, C)$, composable isomorphisms, and notice that $f^{-1} g^{-1} \in \operatorname{Mor}_{C}(C, A)$ is such that $\left(f^{-1} g^{-1}\right)(g f)=\operatorname{id}_{A}$ and $(g f)\left(f^{-1} g^{-1}\right)=\operatorname{id}_{C}$, hence we conclude that $g f$ is an isomorphism and therefore $g f \in \operatorname{Mor}_{G}(A, C)$.

Definition 1.2.20 (Full subcategory). A subcategory D of C (see Definition 1.2.18) is said to be a full subcategory if for all $x, y \in \mathrm{D}$ we have $\operatorname{Mor}_{\mathrm{D}}(x, y)=\operatorname{Mor}_{\mathrm{C}}(x, y)$. The full subcategory D is said to be saturated if $x \in \mathrm{C}$ is also an object of D whenever there exists an object $u \in \mathrm{D}$ such that $x \simeq u$ in $C$.

Definition 1.2.21 (Skeleton). Given a category C, we define the skeleton of C, denoted by sk $C$, to be the full subcategory of $C$ such that, for every $c \in C$, there exists a unique object $s \in \operatorname{sk} C$ such that $c \simeq s$ in $C$. Furthermore, we say that a category $D$ is skeletal if D $=\mathrm{sk} \mathrm{D}$.

Example 1.2.22 (Set based categories). The following are some important categories regarding sets:

- The category Set consists of $\mathcal{U}$-sets and set-maps between such objects.
- The category consisting of finite $\mathcal{U}$-sets and set-maps between them is a full subcategory of Set, we denote it by FinSet.
- We also define the category pSet of pointed $\mathcal{U}$-sets, that is, objects are pairs $(X, x)$, where $X$ is a $\mathcal{U}$-set and $x \in X$. Morphisms $f:(X, x) \rightarrow(Y, y)$ are defined to be maps $f: X \rightarrow Y$ such that $f(x)=y$.

Remark 1.2.23. The structure consisting of all sets and the set-maps between them does not shape a category, since the collection of all sets is not itself a set.

Example 1.2.24. Let $(I, \leqslant)$ be an ordered set. We define a category I associated with $I$ to consist of the collection of objects contained in $I$, and

$$
\operatorname{Mor}_{\mathrm{I}}(i, j)= \begin{cases}\{*\}, & \text { if } i \leqslant j \\ \varnothing, & \text { otherwise }\end{cases}
$$

Definition 1.2.25 (Morphism category). Let C be a category. We'll denote by $\operatorname{Mor}(\mathrm{C})$ the category whose objects are morphisms in C and whose morphisms between given objects $f: x \rightarrow y$ and $g: z \rightarrow w$ are pairs of morphisms $(u, v)$, with $u: x \rightarrow z$ and $v: y \rightarrow w$ such that the following diagram commutes


That is, $\operatorname{Mor}_{\operatorname{Mor}(C)}(f, g):=\left\{(u, v) \in \operatorname{Mor}_{C}(x, z) \times \operatorname{Mor}_{C}(y, w): g u=v f\right\}$.
Proposition 1.2.26 (Slice category). Given a category $C$ and an object $c \in C$. The following define categories:
(SC1) (Slice under $c$ ) A category $c / C$, called slice category of $C$ under $c$, whose objects are morphisms $f \in \operatorname{Mor}_{C}(c, *)$. Given objects $f, g \in c / C$ such that $f: c \rightarrow x$ and $g: c \rightarrow y$, we define a morphism $f \rightarrow g$ as a map $h: x \rightarrow y$ such that the following diagram commutes

that is, $g=h f$.
(SC2) (Slice over c) A category C/c, called the slice category of C over $c$, whose objects are morphisms $f \in \operatorname{Mor}_{C}(*, c)$. Morphisms between objects $f, g \in C / c$ such that $f: x \rightarrow c$ and $g: y \rightarrow c$ are maps $h: x \rightarrow y$ such that the following diagram commutes

so that $f=g h$.
Proof. (SC1) Given objects $f, g \in c / C$ we have from construction that

$$
\operatorname{Mor}_{c / c}(f, g)=\operatorname{Mor}_{c}(\operatorname{cod} f, \operatorname{cod} g),
$$

hence we are ensured of the existence of such morphisms between objects. Given an object $f: c \rightarrow x$, the morphism $\operatorname{id}_{x} \in \operatorname{Mor}(\mathrm{C})$ is such that $f=\operatorname{id}_{x} f$, so that $\mathrm{id}_{x}$ is the identity morphism for $f$. Let $f, g, u \in c / C$ be objects such that $f: c \rightarrow x, g: c \rightarrow$ $y, u: c \rightarrow z$, then there exists morphisms $h \in \operatorname{Mor}_{c / c}(f, u)$ and $\ell: \operatorname{Mor}_{c / c}(u, g)$ so that the following diagram commutes

so that we have a uniquely defined morphism $\ell h \in \operatorname{Mor}_{c}(x, y)$, where $\ell h \in \operatorname{Mor}_{c / c}(f, g)$ and thus this defines a map $\operatorname{Mor}_{c / \mathrm{c}}(f, u) \times \operatorname{Mor}_{c / \mathrm{c}}(u, g) \rightarrow \operatorname{Mor}_{c / \mathrm{c}}(f, g)$. Since $\mathrm{id}_{x}, \mathrm{id}_{z} \in$ $\operatorname{Mor}(c / \mathrm{C})$ for any $x, y \in \mathrm{C}$, then given $f, u \in c / \mathrm{C}$ just as above, the morphism $h: x \rightarrow z$ is such that $\mathrm{id}_{z} h=h=h \mathrm{id}_{x}$. In addition to the objects and morphisms named above, define $v: c \rightarrow w$ and the corresponding morphism $t \in \operatorname{Mor}_{c / c}(g, v)$ so that $t: y \rightarrow w$ and $v=t g$. From the fact that C is a category, we find that $(t \ell) h=t(\ell h)$ and hence the same is true for $c / C$.
(SC2) Let $f, g \in \mathrm{C} / c$, then from definition we have $\operatorname{Mor}_{\mathrm{C} / c}(f, g)=\operatorname{Mor}_{\mathrm{C}}(\operatorname{dom} f, \operatorname{dom} g)$, which is well defined on C . Given an object $f: x \rightarrow c$, there exists $\mathrm{id}_{x}: x \rightarrow x$ so that $f=f \mathrm{id}_{x}$ and hence $\operatorname{id}_{x}$ is the identity morphism for $f$. Define objects $f, g, u \in \mathrm{C} / c$ such that $f: x \rightarrow c, g: y \rightarrow c, u: c \rightarrow z$, and morphisms $h \in \operatorname{Mor}_{c / c}(u, f)$ and $\ell \in \operatorname{Mor}_{c / c}(g, u)$. Then, the following diagram commutes

hence the morphism $h \ell \in \operatorname{Mor}_{c}(y, x)$ defines a morphism $h \ell \in \operatorname{Mor}_{c / c}(g, f)$ so that we can construct a map well defined map $\operatorname{Mor}_{c / c}(g, u) \times \operatorname{Mor}_{c / c}(u, f) \rightarrow \operatorname{Mor}_{c / c}(g, f)$. Moreover, given $f$ and $u$ as above, we have that there exists $\operatorname{id}_{x}, \operatorname{id}_{z} \in \operatorname{Mor}(\mathrm{C} / c)$ so that from the category C it follows that $\mathrm{id}_{x} h=h \mathrm{id}_{z}$. In addition to the above, define $v: w \rightarrow c$ and the morphism $t \in \operatorname{Mor}_{c / c}(v, g)$ so that $t: w \rightarrow y$. Then we have $g=t v$ and since C is a category, we find that $(h \ell) t=h(\ell t)$.

## Initial \& Final Objects

Definition 1.2.27 (Initial, final and zero objects). Given a category C, we define the following objects:
(a) An object $I \in \mathrm{C}$ is said to be an initial object in the category C if, for all $A \in \mathrm{C}$, there exists exactly one morphism $f \in \operatorname{Mor}(I, A)$ so that $\operatorname{Mor}(I, A)=f$.
(b) An object $F \in \mathrm{C}$ is said to be a final object in C if there is exactly one morphism $g \in \operatorname{Mor}(A, F)$ for all given $A \in \mathrm{C}$ and thus $\operatorname{Mor}(A, F)=g$.
(c) An object $O \in \mathrm{C}$ is said to be a zero object if it is both initial and terminal.

Proposition 1.2.28. Let a category $C$, then initial and final objects are said to be unique up to a unique isomorphism:
I. If $I, I^{\prime} \in C$ are initial objects of the category, then $I \simeq I^{\prime}$, where the isomorphism $\varphi_{I}: I \xrightarrow{\leftrightharpoons} I^{\prime}$ is unique.
II. If $F, F^{\prime} \in C$ are initial objects of the category, then $F \simeq F^{\prime}$, where the isomorphism $\varphi_{F}: F \xrightarrow{\leftrightharpoons} F^{\prime}$ is unique.

Proof. Since $I, I^{\prime}$ are both initial objects of $C$ then exists a unique morphism $f \in$ $\operatorname{Mor}\left(I, I^{\prime}\right)$ and there exists a unique $g \in \operatorname{Mor}\left(I^{\prime}, I\right)$ from which we can compose and conclude that $f g=\operatorname{id}_{I^{\prime}}$ and $g f=\operatorname{id}_{I}$ and thus $f$ and $g$ are isomorphisms and also $f^{-1}=g$ and $g^{-1}=f$. For final objects the same reasoning works and thus the proof will be omitted.

We normally say that a given construction satisfies a universal property if such construction is a terminal object of the category.

### 1.3 Duality

Definition 1.3.1 (Opposite category). Let C be a category. We define the opposite category of C as $\mathrm{C}^{\text {op }}$, such that
$(\mathrm{COP} 1) \mathrm{Obj}\left(\mathrm{C}^{\mathrm{Op}}\right)=\mathrm{Obj}(\mathrm{C})$.
(COP2) Given a morphism $f: x \rightarrow y$ in $C$, there exists a corresponding morphism $f^{\mathrm{op}}: y \rightarrow x$ in the $\mathrm{C}^{\mathrm{op}}$. That is $\operatorname{dom} f=\operatorname{cod} f^{\text {op }}$ and $\operatorname{cod} f=\operatorname{dom} f^{\text {op, }}$, these form all of the morphisms in the category $\mathrm{C}^{\mathrm{op}}$.
(COP3) For all $A \in \mathrm{C}^{\mathrm{op}}$, there exists an identity morphism $\mathrm{id}_{A}^{\mathrm{op}} \in \operatorname{End}_{\mathrm{C}_{\mathrm{op}}(A) \text {. }}$
(COP4) A pair of morphisms $f^{\mathrm{op}}, g^{\mathrm{op}} \in \mathrm{C}^{\mathrm{op}}$ is composable, so that $\operatorname{dom} g^{\mathrm{op}}=\operatorname{cod} f^{\mathrm{op}}$, if and only if the pair $g, f \in \mathrm{C}$ is composable, that is $\operatorname{dom} f=\operatorname{cod} g$. Moreover, we define their composition as $g^{\mathrm{op}} f^{\mathrm{op}}=(f g)^{\mathrm{op}}$.

Lemma 1.3.2. The following propositions are equivalent
(a) $f: x \stackrel{\simeq}{\Longrightarrow} y$ is an isomorphism in $C$.
(b) For all $c \in C$, the map

$$
f_{*}: \operatorname{Mor}_{\mathrm{c}}(c, x) \rightarrow \operatorname{Mor}_{\mathrm{c}}(c, y), \quad g \stackrel{f_{*}}{\mapsto} f g
$$

is a bijection.
(c) For all $c \in C$, the map

$$
f^{*}: \operatorname{Mor}_{\mathrm{C}}(x, c) \rightarrow \operatorname{Mor}_{\mathrm{C}}(y, c), \quad g \stackrel{f^{*}}{\longmapsto} g f
$$

is a bijection.
Proof. $((\mathrm{a}) \Rightarrow(\mathrm{b}))$ Let $f: x \xrightarrow{\simeq} y$ be an isomorphism and $c \in \mathrm{C}$ be an object and define $\ell: y \xrightarrow{\simeq} x$ to be its inverse. Given $c \in C$, define $\ell_{*}: \operatorname{Mor}_{\mathrm{C}}(c, y) \rightarrow \operatorname{Mor}_{\mathrm{C}}(c, x)$. Notice that $f_{*} \ell_{*}: \operatorname{Mor}_{C}(c, y) \rightarrow \operatorname{Mor}_{C}(c, y)$ mapping $g \mapsto f_{*}\left(\ell_{*}(g)\right)=f \ell g=g$, hence $f_{*} \ell_{*}=\operatorname{id}_{\text {Morcc }}(c, y)$. Moreover, we have $\ell_{*} f_{*}: \operatorname{Mor}_{C}(c, x) \rightarrow \operatorname{Mor}_{C}(c, x)$ mapping $h \mapsto \ell_{*}\left(f_{*}(h)\right)=h f \ell=h$, hence $\ell_{*} f_{*}=\operatorname{id}_{\operatorname{Mor}_{c}(c, x)}$. This shows that $\ell_{*}$ is the inverse of $f_{*}$ and therefore $f_{*}$ is an isomorphism. $((\mathrm{b}) \Rightarrow(\mathrm{a}))$ Suppose the contrary, so that $f_{*}$ is an isomorphism. In particular, we can take $c=y$ so that $f_{*}: \operatorname{Mor}_{c}(y, x) \rightarrow \operatorname{Mor}_{c}(y, y)$. From the isomorphism property, there exists $\ell \in \operatorname{Mor}_{C}(y, x)$ such that $f_{*}(\ell)=f \ell=\operatorname{id}_{y}$. Consider now that $c=x$, then $f_{*}: \operatorname{Mor}_{c}(x, x) \rightarrow \operatorname{Mor}_{C}(x, y)$. Notice that $f_{*}(\ell f)=f \ell f=\operatorname{id}_{y} f=f$ and $f_{*}\left(\mathrm{id}_{x}\right)=f \mathrm{id}_{x}=f$. Since $f_{*}$ is supposed to be an isomorphism, it follows that $\mathrm{id}_{x}=\ell f$. With this we conclude that $\ell$ is the inverse of $f$ and hence $f$ is an isomorphism.
$((\mathrm{a}) \Leftrightarrow(\mathrm{c}))$ Suppose that $f^{\mathrm{Op}}: y \rightarrow x \in \operatorname{Mor}\left(\mathrm{C}^{\mathrm{O}}\right)$, then from the last paragraph we have that $f^{\mathrm{op}}$ is an isomorphism if and only if $\left(f^{\mathrm{op}}\right)_{*}: \operatorname{Mor}_{c^{\text {op }}}(c, y) \rightarrow \operatorname{Mor}_{c^{\text {op }}}(c, x)$ is an isomorphism. Therefore the dual of such statement is that $f: x \rightarrow y \in \operatorname{Mor}(\mathrm{C})$ is an isomorphism if and only if $\left(f^{\mathrm{op}}\right)_{*}^{\mathrm{op}}=f^{*}: \operatorname{Mor}_{\mathrm{C}}(x, c) \rightarrow \operatorname{Mor}_{\mathrm{C}}(y, c)$ is an isomorphism, since $\operatorname{Mor}_{\text {Cop }}\left(*, *^{\prime}\right)=\operatorname{Mor}_{C}\left(*^{\prime}, *\right)$.

Definition 1.3.3 (Monomorphism). Let a category C. We say that $f \in \operatorname{Mor}_{\mathrm{C}}(x, y)$ is a monomorphism if for all $c \in \mathrm{C}$, and for all $\alpha, \beta \in \operatorname{Hom}(c, x)$ we have that $f \alpha=f \beta$ implies $\alpha=\beta$. Equivalently, for all $c \in \mathrm{C}$ the map $f_{*}: \operatorname{Mor}_{C}(c, x) \rightarrow \operatorname{Mor}_{C}(c, y)$ is injective.

Definition 1.3.4 (Epimorphism). Let a category C. We say that morphism $g \in \operatorname{Hom}_{C}(x, y)$ to be an epimorphism if for all $c \in \mathbb{C}$, and for all $\gamma, \delta \in \operatorname{Hom}_{C}(y, c)$ we have that $\gamma g=\delta g$ implies $\gamma=\delta$. Equivalently, for all $c \in C$ the map $g^{*}: \operatorname{Mor}_{C}(y, c) \rightarrow \operatorname{Mor}_{C}(x, c)$ is injective.

Proposition 1.3.5. Let C be a category. The following are properties regarding monomorphisms and epimorphisms in C:
(a) Every identity morphism is a monomorphism and an epimorphism
(b) The composite of two monomorphisms (or epimorphisms) is a monomorphism (or epimorphism).
(c) If the composition $k f$ is a monomorphism, then $f$ is a monomorphism. Conversely, if $f k$ is an epimorphism, then $f$ is an epimorphism.

Proof. We prove the assertions about monomorphisms, the respective ones for epimorphisms are dually true from the former.
(a) Identities are isomorphism, so clearly they are monomorphisms and epimorphisms.
(b) Let $f$ and $g$ be composable monomorphisms, then if $g f p=g f q$ for two given morphisms $p$ and $q$, for since $g$ is a monomorphism then $f p=f q$ - then using the fact that $f$ is a monomorphism we obtain $p=q$.
(c) Suppose $k f$ is a monomorphism and consider morphisms $g$ and $h$ such that $f g=$ $f h$. Composing with $k$ one sees that $k f g=k f h$ but since $k f$ is a monomorphism, it follows that $g=h$.

Definition 1.3.6. Let $x \xrightarrow{s} y \xrightarrow{r} x$ be morphisms such that $r s=\mathrm{id}_{x}$. We define the following terms
(a) $s$ is said to be a section of $r$. The morphism $s$ is always a monomorphism, being called a split monomorphism.
(b) $r$ is said to be the retraction of $s$. The morphism $r$ is always an epimorphism, being called a split epimorphism.
(c) $x$ is the retract of $y$.

Proposition 1.3.7. A morphism $f \in \operatorname{Mor}_{C}(x, y)$ is a split epimorphism if and only if for all $c \in C$ the map $f_{*}: \operatorname{Mor}_{C}(c, x) \rightarrow \operatorname{Mor}_{C}(c, y)$ is surjective. Dually, $f$ is a split monomorphism if and only if for all $c \in C$ the map $f^{*}: \operatorname{Mor}_{C}(x, c) \rightarrow \operatorname{Mor}_{C}(y, c)$ is surjective.

Proof. $(\Rightarrow)$ Suppose $f: x \rightarrow y$ is a split epimorphism and define $g: y \rightarrow x$ as a section of $f$, that is $f g=\operatorname{id}_{y}$. Let $c \in C$, and $\alpha \in \operatorname{Mor}_{C}(c, y)$ be any morphism. Notice that $g \alpha \in \operatorname{Mor}_{C}(c, x)$, hence we find that $f_{*}(g \alpha)=f g \alpha=\alpha$ and therefore $f_{*}$ is surjective. $(\Leftarrow)$ Suppose $f_{*}$ is surjective, then in particular for $c=y$ we have that $\operatorname{id}_{y} \in \operatorname{im} f_{*}$ and hence $f$ is a split epimorphism. (Dual) Let $f^{\mathrm{op}}: y \rightarrow x$, then from the above proposition $f^{\mathrm{Op}}$ is a split epimorphism if and only if $\left(f^{\mathrm{op}}\right)_{*}: \operatorname{Mor}_{C^{\circ p}}(y, c) \rightarrow \operatorname{Mor}_{C^{\circ p}}(x, c)$ is surjective. Dually we have that $f$ is a monomorphism if and only if $\left(f^{\mathrm{op}}\right)_{*}^{\mathrm{op}}=f^{*}: \operatorname{Mor}_{c}(c, x) \rightarrow \operatorname{Mor}_{C}(c, y)$, which proves the last part.

Proposition 1.3.8. Let $f \in \operatorname{Mor}(\mathrm{C})$. If $f$ is a monomorphism and also a split epimorphism, then $f$ is an isomorphism. Dually, if $f$ is an epimorphism and a split monomorphism, then $f$ is an isomorphism.

Proof. Suppose $f: x \rightarrow y$ is a split epimorphism and $g: y \rightarrow x$ be a section of $f$, then $f g=\operatorname{id}_{y}$ and, moreover $f g f=\operatorname{id}_{y} f=f$ hence from if $f$ is a monomorphism we conclude that $g f=\mathrm{id}_{x}$. Thus $g$ is the inverse of $f$ and hence $f$ is an isomorphism. Dually, let $f^{\mathrm{OP}}: y \rightarrow x$ be a monomorphism and split epimorphism, then $f^{\mathrm{OP}}$ is an isomorphism, which dually means that $f$ is an epimorphism and split monomorphism, then $f$ is an isomorphism.

Lemma 1.3.9. Let $f: x \rightarrow y$ and $g: y \rightarrow z$. The following propositions hold
(a). If $f$ and $g$ are monomorphisms, so is $g f: x \mapsto z$. For the dual proposition, if $f$ and $g$ are epimorphisms, then $g f: x \rightarrow z$ is an epimorphism.
(b). If $g f: x \mapsto z$ is a monomorphism, then $f$ is also monomorphism. Dually, if $g f: x \rightarrow z$ is an epimorphism, then $g$ is an epimorphism.

Proof. (a) Suppose that $f, g$ are both monomorphisms. Given any $c \in C$, let $\alpha, \beta \in$ $\operatorname{Mor}_{C}(c, x)$ be such that $g f \alpha=g f \beta$. In particular, since $g$ is monic, then $f \alpha=f \beta$, but $f$ is also monic, hence $\alpha=\beta$. For the dual part, suppose that $f^{\mathrm{op}}: y \rightarrow x$ and $g^{\mathrm{op}: z \rightarrow y}$ be monic, then from above we have $f^{\mathrm{op}} g^{\mathrm{op}}=(g f)^{\mathrm{op}}$ monic, which dually implies that $g f$ is epic.
(b) Let $g f$ be monic. Then given $c \in C$ and $\alpha, \beta \in \operatorname{Mor}_{C}(c, x)$ such that $f \alpha=f \beta$, then in particular $g(f \alpha)=g(f \beta)$, then from the monic property we have $\alpha=\beta$. Dually, suppose that $(g f)^{\mathrm{Op}}=f^{\mathrm{op}} g^{\mathrm{op}} \in \operatorname{Mor}\left(\mathrm{C}^{\mathrm{OP}}\right)$ is monic, then from above we find that $g^{\mathrm{op}}$ is monic, which dually implies that $g$ is epic.

### 1.4 Functors

Definition 1.4.1 (Covariant functor). Let C and D be categories. A covariant functor $F: C \rightarrow D$ has the following data:
(DF1) For all $c \in C$ exists a corresponding $F c \in D^{4}$.
(DF2) For all $f: c \rightarrow c^{\prime} \in \operatorname{Mor}(\mathrm{C})$ there exists a morphism $F f: F c \rightarrow F c^{\prime} \in \operatorname{Mor}(\mathrm{D})$.
Such data satisfies the two following axioms:
(AF1) For all composable $f, g \in \operatorname{Mor}(\mathrm{C})$, we have $F g \circ F f=F(g \circ f)$.
(AF2) For all $c \in C$ we have $F \mathrm{id}_{c}=\mathrm{id}_{F c} \in \operatorname{Mor}(\mathrm{D})$.

Definition 1.4.2 (Contravariant functor). A contravariant functor from categories $C$ to $D$ is a functor $F: C^{\mathrm{op}} \rightarrow \mathrm{D}$ together with the following data:
(DCF1) For all $c \in C$ exists $F c \in \mathrm{D}$.
(DCF2) For all $f: c \rightarrow c^{\prime} \in \operatorname{Mor}(\mathrm{C})$ we have $F f: F c^{\prime} \rightarrow F c \in \operatorname{Mor}(\mathrm{D})$.

Moreover, a contravariant functor satisfies the following axioms:
(ACF1) For all composable $f, g \in \operatorname{Mor}(\mathrm{C})$ we have $F f \circ F g=F(g \circ f)$.
(ACF2) For all $c \in C$ we have $F \mathrm{id}_{c}=\mathrm{id}_{F c}$.

[^2]This can all be comprised diagrammatically as:


Definition 1.4.3 (Composition of functors). Let $F: C \rightarrow \mathrm{D}$ and $\mathrm{G}: \mathrm{D} \rightarrow \mathrm{B}$ be functors. We define their composition $G F: C \rightarrow B$ by $(G F)(x)=G(F(x))$, for all $x \in C$, and $(G F)(f)=G(F(f))$ for all morphism $f \in \operatorname{Mor}(C)$.

Example 1.4.4. An important example of contravariant functor is op: $C \rightarrow C^{o p}$, where C is some category, defined by the identity on objects and morphisms.

Definition 1.4.5 (Forgetful functor). A functor is said to be forgetful if the functor "forgets" some object, structure or property of its domain category.

Example 1.4.6. We have some classical examples of forgetful functors, for instance, the following are functors that forget the structure of their domain categories:

- The functor $G:$ Grp $\rightarrow$ Set mapping groups to its corresponding underlying set.
- The functor $T:$ Top $\rightarrow$ Set maps any topological space to its corresponding set of points.
- The functor $V, E:$ Graph $\rightarrow$ Set maps the vertices and edges of a graph to the set of such vertices an edges.

Example 1.4.7 (Top ${ }^{\mathrm{op}} \rightarrow$ Ring). Let $C:$ Top $^{\mathrm{op}} \rightarrow$ Ring be a contravariant functor such that for all $X \in$ Top, let $C X$ be the ring of continuous functions $X \rightarrow \mathbf{R}$. The ring operations on $C X$ are defined pointwise, that is, given $p, q: X \rightarrow \mathbf{R} \in C X$ we have $(p \cdot q)(x)=p(x) \cdot q(x)$ and $(p+q)(x)=p(x)+q(x)$ for all $x \in X$. Moreover, given a morphism $f: X \rightarrow Y \in \operatorname{Mor}(T o p)$ we define $C f: C Y \rightarrow C X$ as the composition $(C f)(q)=q f \in \operatorname{Mor}(C X)$ for all $q \in C Y$, that is


We now show that the axioms for the contravariant functor are satisfied by $C$. Let $f: X \rightarrow Y$ and $g: Y \rightarrow Z$, then given any $p \in C Z$ we have

hence $C(f) C(g)=C(g f)$. Moreover, given any $X \in$ Top $^{\text {op }}$ we find that $C \operatorname{id}_{X}: C X \rightarrow$ $C X$ is such that for all $q \in C X, C \operatorname{id}_{X}(q)=q \operatorname{id}_{X}=q$ hence $C \operatorname{id}_{X}=\operatorname{id}_{C X}$. This finishes the proof that $C:$ Top ${ }^{\mathrm{op}} \rightarrow$ Ring is a contravariant functor.

Definition 1.4.8 (Presheaf). Let C be a $\mathcal{U}$-small category. A contravariant functor $\mathrm{C}^{\text {op }} \rightarrow$ Set is called a presheaf on $C$.

Example 1.4.9 $\left(O(X)^{\mathrm{op}} \rightarrow\right.$ Set). Let $X \in$ Top we define $O(X)$ to be the poset category whose objects are open sets of $X$. That is, for sets $U, U^{\prime} \in O(X)$, if $U \subseteq U^{\prime}$, then there exists a morphism $U \rightarrow U^{\prime}$ in $\operatorname{Mor}(O(X))$. A presheaf on the category $O(X)$ is a functor $F: O(X)^{\mathrm{op}} \rightarrow$ Set that assigns $F U=\{f: U \rightarrow \mathbf{R}: f$ continuous $\}$ for all $U \in O(X)$. Moreover, for maps $g: U \rightarrow U^{\prime}$ (that is $U \subseteq U^{\prime}$ ) we have $F g: F U^{\prime} \rightarrow F U$ such that $F g(f)=\left.f\right|_{U}: U \rightarrow \mathbf{R}$ for all $f: U^{\prime} \rightarrow \mathbf{R}$ continuous. Since the restriction of a continuous map is continuous, then $\left.f\right|_{U} \in F U$

Example 1.4.10 (Simplex category). The simplex category $\Delta$ comprises objects that are finite non-empty ordinals and order-preserving morphisms. Simplicial sets are defined as presheaves $\Delta^{\mathrm{op}} \rightarrow$ Set.

Lemma 1.4.11. Functors preserve isomorphisms. Let $C$ and $D$ be categories and $F: C \rightarrow D$ be a functor. Given an isomorphism $f: c \xrightarrow{\simeq} c^{\prime} \in \operatorname{Mor}(\mathrm{C})$, we have that $F f: F c \stackrel{\simeq}{\rightrightarrows} c^{\prime}$ is an isomorphism.

Proof. Denote by $f^{-1}: c^{\prime} \xrightarrow{\simeq} c$ the inverse of $f$. By the composition axiom we have

$$
\begin{gathered}
F\left(f^{-1}\right) F(f)=F\left(f^{-1} f\right)=F \mathrm{id}_{c}=\mathrm{id}_{F c} \\
F(f) F\left(f^{-1}\right)=F\left(f f^{-1}\right)=F \mathrm{id}_{c^{\prime}}=\mathrm{id}_{F c^{\prime}}
\end{gathered}
$$

This shows that $F f^{-1}$ is the right and left inverse of $F f$, hence $F f: F c \stackrel{\simeq}{\leftrightharpoons} F c^{\prime}$ is indeed an isomorphism.

Example 1.4.12 (Group action). Let $G$ be any group and consider the category $B G$ generated by $G$ - that is, $B G$ consists of a unique object * and the morphisms of the category are automorphisms $* \rightarrow *$ given by the elements of $G$. Given a category $C$, a functor $X: B G \rightarrow C-$ given by mapping $X *:=X \in C$ and each object $g \in G$ to an endomorphism $X g:=g_{*}: X \rightarrow X$ - is said to define a left group action on the object $X \in C$. Moreover, the functor $X$ has to obey

- Composition preserving: for any $h, g \in G$, we have $(h g)_{*}=h_{*} g_{*}$.
- Identity: $e_{*}=\mathrm{id}_{X}$.

Since functors preserve isomorphisms and every morphism of $B G$ is an automorphism, it follows that, for all $g \in G$, the map $g_{*}: X \rightarrow X$ is an automorphism - in particular, this implies that $\left(g^{-1}\right)_{*}=g_{*}^{-1}$.

Some particular cases of interest are the following:

- If $\mathrm{C}=$ Set, then the set $X$ together with the actions $\left\{g_{*}: g \in G\right\}$ is called a $G$-set.
- If $\mathrm{C}=\operatorname{Vect}_{k}$, then the $k$-vector space $X$ together with the actions generated by $G$ is said to be a $G$-representation.
- If $\mathrm{C}=\mathrm{Top}$, then the topological space $X$ endowed with the actions generated by $G$ is called a $G$-space.

A right group action is nothing more than a contravariant functor $X: B G^{\mathrm{op}} \rightarrow \mathrm{C}$ such that $X *:=X$ and $X g:=g^{*}: X \rightarrow X$ are endomorphisms. The rules for such a functor are the contravariant preservation of compositions, that is, $(h g)_{*}=g_{*} h_{*}$, and that $e^{*}=\mathrm{id}_{X}$ as before.

Example 1.4.13 (Skeletal functor). Let C be a category. The inclusion functor $J: \mathrm{sk} \mathrm{C} \rightarrow \mathrm{C}$ equivalence of categories, indeed, one can define a quasi-inverse functor $F: C \rightarrow \mathrm{skC}$ by mapping $a \in C$ to the unique object $F a \in \mathrm{sk} \mathrm{C}$ for which there exists an isomorphism $a \simeq F a$. Choosing a collection $\left(\alpha_{a}: a \xrightarrow{\simeq} F a\right)_{a \in \mathrm{C}}$ of isomorphisms where $\alpha_{a}=\operatorname{id}_{a}$ whenever $F a=a$, we can map each morphism $f: a \rightarrow b$ of C to the morphism $F f:=$ $\alpha_{b} f \alpha_{a}^{-1}: F a \rightarrow F b$. From this construction one gets $F J=$ id and the collection $\left(\alpha_{a}\right)_{a \in \mathrm{~A}}$ of chosen isomorphisms define a natural isomorphism $\alpha$ : id $\xlongequal{\Longrightarrow} J F$.

Lemma 1.4.14. Functors preserve split monomorphisms and split epimorphisms.
Proof. Let C and D be categories and consider a functor $F: \mathrm{C} \rightarrow \mathrm{D}$. Define morphisms $x \xrightarrow{s} y \xrightarrow{r} x$ in $\operatorname{Mor}(\mathrm{C})$ such that $r s=\mathrm{id}_{x}$, that is, $s$ is a split monomorphism and $r$ is a split epimorphism. Consider the morphisms Fs: Fx $\rightarrow F y$ and $F r: F y \rightarrow F x$ in $\operatorname{Mor}(\mathrm{D})$. Notice that $F(s) F(r)=F(s r)=F\left(\mathrm{id}_{x}\right)=\mathrm{id}_{F x}$. Hence $F s$ is a split monomorphism and $F r$ is a split epimorphism.

Definition 1.4.15 (Mor functors). Let C be a $\mathcal{U}$-category. Given any $c \in C$, there exists a pair of covariant and contravariant functors, $\operatorname{Mor}(c,-)$ and $\operatorname{Mor}(-, c)$, respectively represented by the object $c$. That is:


We now prove that such definition indeed satisfies the axioms for covariant and contravariant functors. Given morphisms $f: x \rightarrow y$ and $g: y \rightarrow z$ in Mor(C), we see that $g_{*} f_{*}=(g f)_{*}$, moreover $f^{*} g^{*}=(f g)^{*}$. Let $x \in \mathrm{C}$ be any object, then $\mathrm{id}_{x *}=\mathrm{id}_{\operatorname{Mor}(c, x)}=\mathrm{id}_{x}^{*}$. This proves that $\operatorname{Mor}(c,-)$ is covariant and $\operatorname{Mor}(-, c)$ is contravariant.

Definition 1.4.16 (Faithful, full \& its friends). Let C and D be categories. A functor $F: C \rightarrow D$ is said to be
(a) Faithful if for all $x, y \in \mathrm{C}$ the map $\operatorname{Mor}_{\mathrm{C}}(x, y) \mapsto \operatorname{Mor}_{\mathrm{D}}(F x, F y)$ is injective.
(b) Full if for all $x, y \in C$ the map $\operatorname{Mor}_{C}(x, y) \rightarrow \operatorname{Mor}_{\mathrm{D}}(F x, F y)$ is surjective.
(c) Fully faithful if for all $x, y \in \mathrm{C}$ the map $\operatorname{Mor}_{\mathrm{C}}(x, y) \xrightarrow{\simeq} \operatorname{Mor}_{\mathrm{D}}(F x, F y)$ is a bijection.
(d) Essentially surjective if for each $y \in D$ there exists $x \in C$ and an isomorphism $F x \xrightarrow{\simeq} y$ in D.
(e) Conservative if, given a morphism $f$ in C , if $F f$ is an isomorphism in D , then $f$ is an isomorphism in C.

Proposition 1.4.17 (Fully faithful functors and isomorphisms). Let $F: \mathrm{C} \rightarrow \mathrm{D}$ be a fully faithful functor. If $F x \simeq F y$ in $D$, for some $x, y \in C$, then $x \simeq y$ in $C$ for a unique isomorphism.

Proof. If $F x \simeq F y$, let $\phi: F x \xrightarrow{\simeq} F y$ be an isomorphism in D. Since $F$ is fully faithful, there exists unique morphisms $f: x \rightarrow y$ and $g: y \rightarrow x$ in $C$ for which $F f=\phi$ and $F g=\phi^{-1}$. In particular, one has

$$
\operatorname{id}_{F x}=\phi^{-1} \phi=F g \circ F f=F(g f)
$$

From the faithfulness of $F$, since $F \mathrm{id}_{x}=\mathrm{id}_{F x}$, then $g f=\mathrm{id}_{x}$. On the other hand, one also has

$$
\mathrm{id}_{F y}=\phi \phi^{-1}=F f \circ F g=F(f g),
$$

therefore, since $F \mathrm{id}_{y}=\mathrm{id}_{F y}$ we obtain $f g=\mathrm{id}_{y}$. We can now finally conclude that $f$ is an isomorphism and its inverse is $g$.

Definition 1.4.18 (Product \& disjoint union categories). Let $I$ be an indexing set and $\left\{\mathrm{C}_{i}\right\}_{i \in I}$ be a collection of categories associated with $I$. We define the following categories:
(a) The product category $\prod_{i \in I} \mathrm{C}_{i}$ consists of objects $\operatorname{Obj}\left(\prod_{i \in I} \mathrm{C}_{i}\right):=\prod_{i \in I} \operatorname{Obj}\left(\mathrm{C}_{i}\right)$, and morphisms $\operatorname{Mor}_{\prod_{i \in I} \mathrm{c}_{i}}\left(\left(x_{i}\right)_{i \in I},\left(y_{i}\right)_{i \in I}\right):=\prod_{i \in I} \operatorname{Mor}_{c_{i}}\left(x_{i}, y_{i}\right)$ between any two objects $\left(x_{i}\right)_{i \in I}$ and $\left(y_{i}\right)_{i \in I}$ in the category. Composable morphisms $\left(f_{i}\right)_{i \in I}$ and $\left(g_{i}\right)_{i \in I}$ have composition defined component-wise - that is, $\left(f_{i}\right)_{i \in I}\left(g_{i}\right)_{i \in I}:=\left(f_{i} g_{i}\right)_{i \in I}$.
(b) The disjoint union category $\coprod_{i \in I} C_{i}$ consists of objects

$$
\operatorname{Obj}\left(\coprod_{i \in I} \mathrm{C}_{i}\right):=\coprod_{i \in I}\left\{(x, i): i \in I \text { and } x \in \mathrm{C}_{i}\right\}
$$

and morphisms

$$
\operatorname{Mor}_{\amalg_{i \in I} c_{i}}((x, i),(y, j)):= \begin{cases}\operatorname{Mor}_{c_{i}}(x, y), & \text { if } i=j \\ \varnothing, & \text { otherwise }\end{cases}
$$

for any objects $(x, i)$ and $(y, j)$ in the category.
Moreover, if $\left\{\mathrm{D}_{i}\right\}_{i \in I}$ is another collection of categories, and $\left\{F_{i}: \mathrm{C}_{i} \rightarrow \mathrm{D}_{i}\right\}_{i \in I}$ is a collection of functors, we associate to each of the above categories the functors $\prod_{i \in I} F_{i}$ and $\coprod_{i \in I} F_{i}$.

Example 1.4.19 (Orbit category). Let $G$ be a group. We define the orbit category associated to $G$ as $O_{G}$ whose objects are subgroups $H \subseteq G$, identified by the left $G$-set $G / H$ of left cosets of $H$. The morphisms $\phi: G / H \rightarrow G / Q$ are maps commuting with the left $G$-action, that is, $\phi\left(g_{*} h\right)=g_{*} \phi(h)$ - such maps are called G-equivariant.

Proposition 1.4.20 (Bifunctor). Let A, B and C be any categories. A functor

$$
F: \mathrm{A} \times \mathrm{B} \longrightarrow \mathrm{C}
$$

is called a bifunctor. The functor $F$ is defined so that, given any objects $x \in \mathrm{~A}$ and $y \in \mathrm{~B}$,

$$
F(x,-): \mathrm{B} \longrightarrow \mathrm{C} \text { and } F(-, y): \mathrm{A} \longrightarrow \mathrm{C}
$$

are both functors. Moreover, given any morphisms $f: x \rightarrow y$ in A and $g: x^{\prime} \rightarrow y^{\prime}$ in B , the following diagram commutes


Notice that the product of small categories can also be understood as a bifunctor

$$
\times \text { Cat } \times \text { Cat } \longrightarrow \text { Cat }
$$

which associates to each pair of small categories $(A, B)$ the product category $A \times B$, and any pair of functors $(F: A \rightarrow B, G: C \rightarrow D)$ is mapped to a bifunctor

$$
F \times G: \mathrm{A} \times \mathrm{C} \rightarrow \mathrm{~B} \times \mathrm{D}
$$

such that $(F \times G)(f, g):=(F f, G g)$ and $(F \times G)(A, C):=(F A, G C)$, for any pair of morphisms $(f, g) \in \operatorname{Mor}(\mathrm{A}) \times \operatorname{Mor}(\mathrm{C})$ and pair of objects $(A, C) \in \mathrm{A} \times \mathrm{C}$.

Definition 1.4.21 (Functor-induced categories). Let $F: C \rightarrow D$ be a functor between categories C and D , and let $y \in \mathrm{D}$. We define the following two categories:
(a) The category $C_{y}$ is defined to consist of objects

$$
\operatorname{Obj}\left(\mathrm{C}_{y}\right):=\{(x, s): x \in \mathrm{C} \text { and } s: F x \rightarrow y \text { in } \mathrm{D}\},
$$

and morphisms between any objects $(a, s),(b, t) \in C_{y}$ are defined to be

$$
\operatorname{Mor}_{C_{y}}((a, s),(b, t)):=\left\{f \in \operatorname{Mor}_{\mathrm{C}}(a, b): s=t F(f) \text { in } \mathrm{D}\right\},
$$

that is, a morphism $f:(a, s) \rightarrow(b, t)$ makes to following diagram commute


Together with such category, we define a faithful functor $j_{y}: \mathrm{C}_{y} \rightarrow \mathrm{C}$ by $j_{y}(x, s):=x$, acting as a projection.
(b) The category ${ }^{y}$ is defined to consist of objects

$$
\operatorname{Obj}\left(\mathrm{C}^{y}\right):=\{(x, s): x \in \mathrm{C} \text { and } s: y \rightarrow F x \text { in } \mathrm{D}\}
$$

and morphisms between any objects $(a, s),(b, t) \in C^{y}$ are defined to be

$$
\operatorname{Mor}_{C^{y}}((a, s),(b, t)):=\left\{f \in \operatorname{Mor}_{C}(a, b): t=F(f) s \text { in } D\right\}
$$

that is, a morphism $f:(a, s) \rightarrow(b, t)$ makes to following diagram commute


Together with such category, we define a faithful functor $j^{y}: \mathrm{C}^{y} \rightarrow \mathrm{C}$ by $j^{y}(x, s):=x$, which acts as a projection.

Definition 1.4.22 (Equivalence classes). Let $C$ be a category and $\sim$ denote an equivalence relation on the objects of $C-$ where $x \sim y$ if $\operatorname{Mor}_{C}(x, y) \neq \varnothing$. We denote the collection of all equivalence classes on the objects of $C$ by $\pi_{0}(C)$.

Corollary 1.4.23. A category $C$ is connected if and only if $\pi_{0}(C)$ consists of a single element.

Proof. If $\pi_{0}(C)$ is a single object, every equivalence class on $C$ is such that every element is equivalent to each other, which implies that the collection of morphisms $\operatorname{Mor}_{\mathrm{C}}(x, y)$, between any two elements $x, y \in \mathrm{C}$, is non-empty - thus clearly C is connected.

Definition 1.4.24 (Isomorphisms of monomorphisms \& epimorphisms). Let C be a category and $x, y, z \in \mathrm{C}$ be any objects. We define the following concepts:
(a) Two monomorphisms $f: x \mapsto z$ and $g: y \mapsto z$ in C are said to be isomorphic in C if there exists an isomorphism $h: x \xrightarrow{\simeq} y$ for which the following diagram commutes in


This is equivalent to $f$ and $g$ being isomorphic in $\mathrm{C}_{z}$.
(b) Two epimorphisms $f: x \rightarrow z$ and $g: y \rightarrow z$ in C are said to be isomorphic in C if there exists an isomorphism $h: x \xrightarrow{\simeq} y$ for which the following diagram commutes in


This is equivalent to $f$ and $g$ being isomorphic in $\mathrm{C}^{z}$.
Definition 1.4.25. Let C be a category and $x \in \mathrm{C}$ be any object. We define the following:
(a) An isomorphism class of a monomorphism with target $x$ is called a subobject of $x$.
(b) An isomorphism class of an epimorphism with source $x$ is called a quotient of $x$.

These isomorphism classes are given in the sense of Definition 1.4.24.
Example 1.4.26 (Ordering subobject). One can order the collection of subobjects of a given object $x \in C$ by defining a relation $[f: y \mapsto x] \leqslant[g: z \mapsto x]$ if there exists $h: y \rightarrow z$ such that


Moreover, if $h$ exists, then it's unique.

### 1.5 Natural Transformations

Definition 1.5.1 (Natural transformation). Let C and D be categories, and consider functors $F, G: C \rightrightarrows$ D. A natural transformation $\alpha: F \Rightarrow G$ consists of morphisms $\alpha_{x}: F x \rightarrow$ $G x$, for all $x \in C$, such that, for any morphism $f: x \rightarrow y$ in $C$, the following diagram commutes


Moreover, if $L: \mathrm{C} \rightarrow \mathrm{D}$ is another functor, and $\beta: G \Rightarrow L$ is a natural transformation, we define the composition of natural transformations $\alpha$ and $\beta$ as the map $\beta \alpha$ :F $\Rightarrow L$ such that $(\beta \alpha)_{x}:=\beta_{x} \alpha_{x}$ for every $x \in \mathrm{C}$.

Definition 1.5.2 (Functor category). Let C and D be categories. We define a category Fct(C, D) whose objects are functors C $\rightarrow$ D, and morphisms are natural transformations between functors.

Proposition 1.5.3 (Horizontal composition). Let $\mathrm{A}, \mathrm{B}$ and C be categories. Consider functors $F, F^{\prime}: \mathrm{A} \rightrightarrows \mathrm{B}$, and $G, G^{\prime}: \mathrm{B} \rightrightarrows \mathrm{C}$. Let $\alpha: F \Rightarrow F^{\prime}$ and $\beta: G \Rightarrow G^{\prime}$ be natural transformations - that is,


There exists an induced natural transformation $\beta * \alpha$ : GF $\Rightarrow G^{\prime} F^{\prime}$ - called the horizontal composition, also called Godement product, of $\alpha$ and $\beta$ - such that, for all $a \in \mathrm{~A}$,

$$
(\beta * \alpha)_{a}=\beta_{F^{\prime} a} G\left(\alpha_{a}\right)=G^{\prime}\left(\alpha_{a}\right) \beta_{F a} .
$$

The horizontal composition can be depicted by the following diagram


Proof. Notice that the naturality of $\beta * \alpha$ solely depends on the naturality of both $\alpha$ and $\beta$. Indeed, for every $a \in \mathrm{~A}$ and morphism $f: a \rightarrow a^{\prime}$ in A the following diagram

is commutative, which proves that $\beta * \alpha$ is a natural transformation.
Notation 1.5.4. For the ease of notation, we denote the vertical composition $\alpha * \operatorname{id}_{F}$ by $\alpha * F$.

Definition 1.5.5 (Vertical composition). Let $C$ and $D$ be categories, and consider functors and natural transformations given in the following diagram


We define the vertical composition of $\beta$ with $\alpha$ as the natural transformation $\beta \circ \alpha: F \Rightarrow G$ - diagrammatically,


Proposition 1.5.6. Consider the following diagram, with categories A, B and C, functors $F, G, H, K, L, M$, and natural transformations $\alpha, \beta, \gamma$ and $\delta$ :


This diagram is such that the following equality holds:

$$
(\delta * \gamma) \circ(\beta * \alpha)=(\delta \circ \beta) *(\gamma \circ \alpha)
$$

Where by o we denote the vertical composition and by * we denote the horizontal composition.

Proof. The composition $(\delta * \gamma) \circ(\beta * \alpha): G F \Rightarrow M L$ is given by the following diagrams


On the other hand, the composition $(\delta \circ \beta) *(\gamma \circ \alpha): G F \Rightarrow M L$ is given by the following diagrams


This is sufficient to prove the assertion.

Notice that, given a functor $\phi: \mathrm{C} \rightarrow \mathrm{D}$ between categories C and D , for every category A, there arises a natural functor

$$
\phi^{*}: \operatorname{Fct}(\mathrm{A}, \mathrm{C}) \longrightarrow \operatorname{Fct}(\mathrm{I}, \mathrm{D}), \text { mapping } F \longmapsto \phi F
$$

Lemma 1.5.7. If $\phi$ is a faithful functor (respectively, fully faithful), then so is the functor $\phi^{*}$ for any category A.

Proof. Given any two functors $F, G$ : A $\rightrightarrows C$, let $\eta: F \Rightarrow G$ be any natural transformation. If we apply $\phi^{*}$, we get the following commutative diagram - for every pair $x, y \in \mathrm{~A}$ and every morphism $f: x \rightarrow y$ in A ,


Since $\phi$ is faithful (or fully faithful), the mappings $\eta_{x} \mapsto \phi \eta_{x}$ and $\eta_{y} \mapsto \phi \eta_{y}$ are both injective (or bijective), thus the natural map

$$
\operatorname{Mor}_{\mathrm{Fct}(\mathrm{~A}, \mathrm{C})}(F, G) \longrightarrow \operatorname{Mor}_{\mathrm{Fct}(\mathrm{~A}, \mathrm{D})}(\phi F, \phi G), \text { mapping } \eta \mapsto \phi^{*} \eta
$$

is ensured to be injective (or bijective).
We consider now the category consisting of $\mathcal{U}$-small categories and the morphisms are functors between them, we denote this category by $\mathcal{U}$-Cat. Notice that, given any two $\mathcal{U}$-small categories C and D , the collection of functors between them also forms a category $\operatorname{Mor}_{\mathcal{U} \text {-Cat }}(\mathrm{C}, \mathrm{D})=\operatorname{Fct}(\mathrm{C}, \mathrm{D})$ - this emergent structure gives birth to the concept of a 2-category.

Definition 1.5.8 (Isomorphism of categories). Let $C$ and $D$ be categories. We say that $C$ is isomorphic to D if there are morphisms $F: \mathrm{C} \rightarrow \mathrm{D}$ and $G: \mathrm{D} \rightarrow \mathrm{C}$ such that $G F=\mathrm{id}_{\mathrm{C}}$, and $F G=\mathrm{id}_{\mathrm{D}}$.

A weaker and even more important concept it that of an equivalence between categories.

Definition 1.5.9 (Equivalence of categories). Let C and D be categories. We say that a functor $F: C \xrightarrow{\simeq} D$ is an equivalence of the categories $C$ and $D$ if there exists a functor $G: D \rightarrow C$, and two natural isomorphisms $\alpha: G F \xlongequal{\Longrightarrow} \mathrm{id}_{\mathrm{C}}$ and $\beta: F G \xlongequal{\Longrightarrow} \mathrm{id}_{\mathrm{D}}$. If this is the case, we say that $F$ and $G$ are quasi-inverses of each other.

Lemma 1.5.10. Let $F: C \xrightarrow{\simeq} \mathrm{D}$ and $\mathrm{G}: \mathrm{D} \xrightarrow{\cong} \mathrm{C}$ be equivalences of given categories C and $D$, and suppose that $F$ and $G$ are quasi-inverses. Then there are natural isomorphisms $\alpha: G F \xlongequal{\Longrightarrow} \mathrm{id}_{\mathrm{C}}$ and $\beta: F G \xlongequal{\Longrightarrow} \mathrm{id}_{\mathrm{D}}$ for which

$$
F \alpha=\beta F \quad \text { and } \quad \alpha G=G \beta
$$

Proof. Let $x \in \mathrm{C}$ be any object. Notice that, $\operatorname{since}_{\mathrm{id}}^{\mathrm{C}} \mathrm{x}=x$ and $^{\mathrm{id}}{ }_{\mathrm{D}} F x=F x$, it follows that

is commutative - thus indeed $F \alpha=\beta F$. Now if we let $y \in D$ be any other object, since $\operatorname{id}_{\mathrm{D}} y=y$ and $\mathrm{id}_{\mathrm{C}} G y=G y$, we get the following commutative diagram

therefore $\alpha G=G \beta$ as wanted.

Lemma 1.5.11. Let $F: C \rightarrow D$ be a functor, and $D_{0}$ be a full subcategory of $D$ such that, for all $x \in C$, there exists $y \in D_{0}$ and an isomorphism $F x \simeq y$.

Denote by $t: \mathrm{D}_{0} \hookrightarrow \mathrm{D}$ the canonical embedding functor. Then there exists a functor $F_{0}: \mathrm{C} \rightarrow \mathrm{D}_{0}$ and a natural isomorphism $\alpha: F \xlongequal{\Longrightarrow} \iota F_{0}$. Moreover, $F_{0}$ is unique up to unique isomorphism ${ }^{5}$. This can be diagrammatically expressed by the following quasicommutative diagram ${ }^{6}$


Proof. We first build the functor $F_{0}: C \rightarrow D_{0}$ :

- Let $x \in \mathrm{C}$ be any object. By means of Zorn's Lemma (see Lemma 5.6.5), we choose $a \in \mathrm{D}_{0}$ for which exists an isomorphism $\phi_{x}: a \stackrel{\simeq}{\leftrightharpoons} F(x)$ in D - and consequently we let $F_{0} x:=a$.
- Given any morphism $f: x \rightarrow y$ in C , we know from construction that the objects $F_{0} x:=a$ and $F_{0} y:=b$ are defined so that there exists isomorphisms $\phi_{x}: a \xrightarrow{\simeq} F x$ and $\phi_{y}: b \xrightarrow{\simeq} F y$ in the category $D$. This allows us to define the morphism $F_{0} f: F_{0} x \rightarrow F_{0} y$ in $D$ as the mapping $F f:=\phi_{y}^{-1}(F f) \phi_{x}$ - that is, so that the

[^3]following diagram commutes
\[

$$
\begin{aligned}
F_{0} x \xrightarrow{F_{0} f} & F_{0} y \\
\phi_{x} \downarrow \simeq & \simeq \phi_{y} \\
F x \xrightarrow[F f]{ } & F y
\end{aligned}
$$
\]

Notice that from this definition we find naturally that the composition condition is met - given any other morphism $g: y \rightarrow z$ in C have that $F_{0}(g f)=\left(F_{0} g\right)\left(F_{0} f\right)$.

This proves the existence of $F_{0}$ as a functor. For the isomorphism $\alpha$, we can define for each pair $x, y \in \mathrm{C}$ the morphisms $\alpha_{x}:=\phi_{x}^{-1}$ and $\alpha_{y}:=\phi_{y}^{-1}$, so that the following diagram commutes


For the uniqueness of $F_{0}$ up to unique isomorphism, let $G: C \rightarrow D_{0}$ be another functor, together with a natural isomorphism $\beta: F \cong \stackrel{\cong}{\Longrightarrow}$. Define $\eta: G \xlongequal{\Longrightarrow} F_{0}$ so that, for each $x \in \mathrm{C}$, we have $\eta_{x}:=\alpha_{x} \beta_{x}^{-1}$ - then, $\eta$ is clearly an isomorphism and also uniquely defined, thus the proposition follows.

Lemma 1.5.12. Let $C$ be any category. There exists a full subcategory $C_{0}$ of $C$ such that the embedding functor $t: \mathrm{C}_{0} \hookrightarrow \mathrm{C}$ is a category equivalence and any two isomorphic elements in $C_{0}$ are equal to each other.

Proof. Let $\sim$ be the equivalence relation on the set $\operatorname{Obj}(\mathrm{C})$ where $x \sim y$ if and only if we have $x \simeq y$ in $C$. By means of Zorn's lemma, choose for each equivalence class a representative - and define $C_{0}$ as the full subcategory of $C$ consisting of such representatives. In this way, if $x, y \in \operatorname{Obj}\left(\mathrm{C}_{0}\right)$ satisfy $x \simeq y$ in C , then necessarily $x=y$.

If we apply Lemma 1.5 .11 to the identity functor $\mathrm{id}_{\mathrm{C}}$, we find the existence of a unique functor $F_{0}: C \rightarrow C_{0}$ and a natural isomorphism $\alpha: \mathrm{id}_{\mathrm{C}} \xlongequal{\Longrightarrow} \iota F_{0}$ for which $\iota F_{0} \simeq \mathrm{id}_{\mathrm{C}}$. Moreover, we have the chain of isomorphisms:

$$
\iota\left(F_{0} \iota\right)=\left(\iota F_{0}\right) \iota \simeq \mathrm{id}_{c} \iota \simeq \iota \simeq \iota \mathrm{id}_{\mathrm{C}_{0}}
$$

From the fact that $\iota$ is fully faithful, we obtain $F_{0} \iota \simeq \mathrm{id}_{\mathrm{C}_{0}}$.
Proposition 1.5.13. A functor $F: C \rightarrow D$ is an equivalence of categories if and only if $F$ is fully faithful and essentially surjective.

Proof. Suppose $F$ is an equivalence of categories, then exists a quasi inverse $G: D \rightarrow C$ and natural isomorphisms $\alpha: G F \cong \operatorname{id}_{\mathrm{C}}$ and $\beta: F G \cong \mathrm{id}_{\mathrm{D}}$. We now prove that $F$ is both fully faithful and essentially surjective.

- (Fully Faithful) Let $g: F x \rightarrow F y$ be any morphism in D, thus if there exists an $f: x \rightarrow y$ in $C$ for which $F f=g$, it must be the case that $f=\alpha_{y}(G F f) \alpha_{x}^{-1}=$ $\alpha_{y}(G g) \alpha_{x}^{-1}$. Indeed, if $f=\alpha_{y}(G g) \alpha_{x}^{-1}: x \rightarrow y$ is a morphism in $\mathbf{C}$, from the distributivity over composition, we have

$$
G F f=G F \alpha_{y} \circ G F(G g) \circ G F \alpha_{x}^{-1}
$$

then by the naturality of $\alpha$ on the map $\alpha_{x}^{-1}$ implies

but since $\alpha_{x}^{-1}$ is an isomorphism, then $G F \alpha_{x}^{-1} \alpha_{x}^{-1}=\alpha_{G F x}^{-1} \alpha_{x}^{-1}$ implies in

$$
G F \alpha_{x}^{-1}=\alpha_{G F x}^{-1} .
$$

On the other hand, the same can be done for $G F \alpha_{y}$, yielding $G F \alpha_{y}=\alpha_{G F y}$, therefore

$$
\begin{equation*}
G F f=\alpha_{G F y}(G F G g) \alpha_{G F x}^{-1} \tag{1.1}
\end{equation*}
$$

The naturality of $\alpha$ on the map $G g$ gives

that is, $G F(G g) \alpha_{G F x}^{-1}=\alpha_{G F y}^{-1} G g$, thus substituting in Eq. (1.1) we obtain finally that

$$
G F f=G g
$$

From this we conclude that $F$ is fully faithful.

- (Essentially surjective) Let $a \in \mathrm{D}$ be any element and simply consider $x=G a \in \mathrm{D}$ then from $\beta$ we obtain that $F x=F G a \simeq a$.

For the converse, suppose $F$ is both fully faithful and essentially surjective. Using Lemma 1.5.12, let $C_{0}$ be the full subcategory of $C$ such that $\iota_{C}: C_{0} \hookrightarrow C$ is an equivalence and if $x \simeq y$ in $C$ then $x=y$ in $C_{0}$. Let $\kappa_{C}$ be a quasi-inverse of $\iota_{C}$. Apply Lemma 1.5.12 again, but now to the category $D$, yielding a category $D_{0}$, an embedding $l_{D}$ and a quasiinverse $\kappa_{D}$. Notice that the composition of functors $\kappa_{D} F t_{C}: C_{0} \xrightarrow{\simeq} D_{0}$ is an isomorphism. Let $H$ be the inverse of $\kappa_{D} F \iota_{C}$, and define $G:=\iota_{C} H \kappa_{D}$, then $G$ is a quasi-inverse of $F$. $\quad$, Corollary 1.5.14. Let $F: C \rightarrow D$ be a fully faithful functor. Then there exists a full subcategory $\mathrm{D}_{0}$ of D and an equivalence of categories $G: C \xrightarrow{\leftrightharpoons} \mathrm{D}_{0}$ for which $F$ is isomorphic to $\iota_{D} G$ - where $\iota_{D}: D_{0} \hookrightarrow D$ is the embedding functor.

Proof. Define $\mathrm{D}_{0}$ to be the full category with $\operatorname{Obj}\left(\mathrm{D}_{0}\right):=\{F x: x \in \mathrm{C}\}$. Now define $G: D \rightarrow D_{0}$ to be the functor $G=F$, we find that clearly $F \simeq \iota_{D} G$.

■
Example 1.5.15. For any two given categories $C$ and $D$, there is an isomorphism of categories $\operatorname{Fct}(C, D)^{\mathrm{op}} \simeq \operatorname{Fct}\left(\mathrm{C}^{\mathrm{op}}, \mathrm{D}^{\mathrm{op}}\right)$. Explicitly, such isomorphism maps $F \mapsto \mathrm{op}$ Fop.

Definition 1.5.16 (Essentially $\mathcal{U}$-small). A category C is said to be essentially $\mathcal{U}$-small if it is equivalent to a $\mathcal{U}$-small category. Equivalently, C is essentially $\mathcal{U}$-small if and only if C is a $\mathcal{U}$-category and there exists a subset $S \subseteq \mathrm{Obj}(\mathrm{C})$ that is $\mathcal{U}$-small for which, given any $x \in C$, there exists $y \in S$ such that $x \simeq y$.
Definition 1.5.17 (Half-full). We define the concepts of half-fullness for functors and subcategories:
(a) A functor $F: \mathrm{C} \rightarrow \mathrm{D}$ is said to be half-full if for any two $x, y \in \mathrm{C}$ for which $F x \simeq F y$ in D , there exists an isomorphism $x \simeq y$ in C .
(b) A subcategory $C_{0}$ of $C$ is said to be half-full if the embedding functor $t: C_{0} \hookrightarrow C$ is half-full.

An analogous but less strict proposition when compared to Corollary 1.5.14 is done by substituting the condition of fully faithfulness to only that of faithful and half-full. It goes as follows.

Proposition 1.5.18. Let $F: C \rightarrow D$ be a faithful and half-full functor. Then there exists a subcategory $D_{0}$ of $D$ for which

$$
F(\operatorname{Obj}(\mathrm{C})) \subseteq \operatorname{Obj}\left(\mathrm{D}_{0}\right), \text { and } F(\operatorname{Mor}(\mathrm{C})) \subseteq \operatorname{Mor}\left(\mathrm{D}_{0}\right)
$$

Moreover, $F$ induces an equivalence of categories $C \simeq D_{0}$ and the embedding $\mathrm{D}_{0} \hookrightarrow \mathrm{D}$ is faithful and half-full.

Proof. Let $\mathrm{D}_{0}$ be the category with $\operatorname{Obj}\left(\mathrm{D}_{0}\right):=\{F(x): x \in \mathrm{C}\}$, and for each two $x, y \in \mathrm{C}$ define $\operatorname{Mor}_{\mathrm{D}_{0}}(F x, F y):=F\left(\operatorname{Mor}_{C}(x, y)\right)$ - so that $\operatorname{Mor}_{\mathrm{D}_{0}}(F x, F y) \subseteq \operatorname{Mor}_{\mathrm{D}}(F x, F y)$. Since $F$ is half-full, the definition of $\operatorname{Mor}_{\mathrm{D}_{0}}(F x, F y)$ is independent of the initial choice of $x, y \in C$ - indeed, if $x^{\prime}, y^{\prime} \in C$ are such that $F x^{\prime} \simeq F x$ and $F y^{\prime} \simeq F y$ in $D_{0}$, then $x^{\prime} \simeq x$ and $y^{\prime} \simeq y$ in $C$, thus $\operatorname{Mor}_{c_{0}}(x, y) \simeq \operatorname{Mor}_{c_{0}}\left(x^{\prime}, y^{\prime}\right)$.

Restricting the codomain of $F$ to $\mathrm{D}_{0}$, we find that the induced functor $F: C \rightarrow \mathrm{D}_{0}$ is fully faithful and essentially surjective - thus by Proposition 1.5.13 $F$ is an equivalence of categories.

### 1.6 Comma Categories

Definition 1.6.1 (Comma category). Let $F: \mathrm{A} \rightarrow \mathrm{C}$ and $G: \mathrm{B} \rightarrow \mathrm{C}$ be two functors. The comma category $F \downarrow G$ induced by the functors $F$ and $G$ is defined as follows:

- The objects of $F \downarrow G$ are triples $(a, f, b)$ - where $a \in \mathrm{~A}, b \in \mathrm{~B}$, and $f: F a \rightarrow G b$ is a morphism of C .
- A morphism $\phi:(a, f, b) \rightarrow\left(a^{\prime}, f^{\prime}, b^{\prime}\right)$ in $F \downarrow G$ is a pair of $(\alpha, \beta)$ where $\alpha: a \rightarrow a^{\prime}$ is a morphism of $A$, while $\beta: b \rightarrow b^{\prime}$ is a morphism of $B$ - moreover, such morphisms are such that the following diagram commutes

- The composition of compatible morphisms $(\alpha, \beta)$ and $\left(\alpha^{\prime}, \beta^{\prime}\right)$ in $F \downarrow G$ is induced by the composition law of A and B as follows:

$$
\left(\alpha^{\prime}, \beta^{\prime}\right) \circ(\alpha, \beta):=\left(\alpha^{\prime} \alpha, \beta^{\prime} \beta\right) .
$$

Proposition 1.6.2 (Projection functors in $F \downarrow G$ ). Let $F: \mathrm{A} \rightarrow \mathrm{C}$ and $G: \mathrm{B} \rightarrow \mathrm{C}$ be functors. There are two functors $A: F \downarrow G \rightarrow \mathrm{~A}$ and $B: F \downarrow G \rightarrow \mathrm{~B}$, and a canonical natural transformation $\eta: F A \Rightarrow G B$. The scenario can be depicted in the following diagram


It is to be noted that such diagram does not commute in general - that is, we may have comma categories where $F A \neq G B$.

Proof. Define the functors $A$ and $B$ as follows - for every object $(a, f, b) \in F \downarrow G$ and morphism $(\alpha, \beta) \in \operatorname{Mor}(F \downarrow G)$ :

- $\operatorname{Map} A(a, f, b):=a$, and $A(\alpha, \beta):=\alpha$.
- Map $B(a, f, b):=b$, and $A(\alpha, \beta):=\beta$.

For the natural transformation, simply define $\eta_{(a, f, b)}:=f: F a \rightarrow F b$ which must be natural from the construction of comma categories.

Proposition 1.6.3 (Comma category property). Let $F: A \rightarrow C$ and $G: B \rightarrow C$ be two functors. If there exists a category D with two functors $A^{\prime}: \mathrm{D} \rightarrow \mathrm{A}$ and $B^{\prime}: \mathrm{D} \rightarrow \mathrm{B}$, and a natural transformation $\eta^{\prime}: F A^{\prime} \Rightarrow G B^{\prime}$ - then there exists a unique functor $W: \mathrm{D} \rightarrow F \downarrow G$ such that

$$
\alpha * W=\alpha^{\prime},
$$

and that the following diagram commutes


Proof. For every $d \in D$, define $W d:=\left(A^{\prime} d, \eta_{d^{\prime}}^{\prime} B^{\prime} d\right)$, and for each morphism $f \in \operatorname{Mor}(\mathrm{D})$, we define $W f:=\left(A^{\prime} f, B^{\prime} f\right)$. This completely determines $W$ and thus shows that, if it exists, it is unique. To show that $W: \mathrm{D} \rightarrow F \downarrow G$ is indeed a functor, we note that:

- Given morphisms $f: x \rightarrow y$ and $g: x \rightarrow z$ of D , we have

$$
\begin{aligned}
W g \circ W f & =\left(A^{\prime} g, B^{\prime} g\right) \circ\left(A^{\prime} f, B^{\prime} f\right) \\
& =\left(A^{\prime} g \circ A^{\prime} f, B^{\prime} g \circ B^{\prime} f\right) \\
& =\left(A^{\prime}(g f), B^{\prime}(g f)\right) \\
& =W(g f) .
\end{aligned}
$$

Note that although we used the same symbol $\circ$ for composition, one should note that they have different laws.

- Moreover, for any $d \in D$ we have

$$
W \mathrm{id}_{d}=\left(A^{\prime} \mathrm{id}_{d}, B^{\prime} \mathrm{id}_{d}\right)=\left(\mathrm{id}_{A^{\prime} d}, \mathrm{id}_{B^{\prime} d}\right)=\mathrm{id}_{W d} .
$$

Definition 1.6.4 (Category of elements). Let $F: C \rightarrow$ Set be a covariant functor. The category of elements of $C$ associated to $F$ is denoted by $\mathrm{El}_{F}(\mathrm{C})$ whose objects are pairs $(c, s)$ for $c \in C$ and $s \in F c$, and morphisms between any two objects $(c, s),\left(c^{\prime}, s^{\prime}\right) \in \mathrm{El}_{F}(\mathrm{C})$ is defined as

$$
\operatorname{Mor}_{\mathrm{El}_{F}(\mathrm{C})}\left((c, s),\left(c^{\prime}, s^{\prime}\right)\right):=\left\{u \in \operatorname{Mor}_{\mathrm{C}}\left(c, c^{\prime}\right): F(u)(s)=s^{\prime}\right\}
$$

For a contravariant $G: C^{o p} \rightarrow$ Set, the category of elements of $C$ associated to $G$ is composed of objects $(c, s)$ for $c \in C$ and $s \in G c$, and the collection of morphisms between objects $(c, s),\left(c^{\prime}, s^{\prime}\right) \in \mathrm{El}_{G}(\mathrm{C})$ is given by

$$
\operatorname{Mor}_{E l_{G}(\mathrm{C})}\left((c, s),\left(c^{\prime}, s^{\prime}\right)\right):=\left\{u \in \operatorname{Mor}_{C}\left(c^{\prime}, c\right): F(u)\left(s^{\prime}\right)=s\right\} .
$$

Corollary 1.6.5. Let $F: C \rightarrow D$ be a functor. The category of elements $\mathrm{El}_{F}(\mathrm{C})$ is exactly the comma category $1 \downarrow F-$ where $1: 1 \rightarrow$ Set is the functor from the discrete category 1 with a single object $\star$ to the category of sets, mapping $\star \mapsto\{*\}$.

### 1.7 Yoneda Lemma

Remark 1.7.1. I again stress that Set, for us, is defined to be the category whose objects are $\mathcal{U}$-sets for a given universe $\mathcal{U}$ and set-functions between these sets - this is a relevant remark, since confusions with that would lead one to undesirable size issues.

Definition 1.7.2 (Category of presheaves \& Yoneda functors). Given a $\mathcal{U}$-category C, we define the big category of presheaves $\mathrm{Psh}(\mathrm{C})$ and a big category of functors $\operatorname{Psh}\left(\mathrm{C}^{\mathrm{Op}}\right)=$ [ $C, \operatorname{Set}$ ]. Together with such categories we define functors

$$
\begin{aligned}
& H_{\mathrm{C}}: \mathrm{C} \longrightarrow \operatorname{Psh}(\mathrm{C}), \text { mapping } x \mapsto \operatorname{Mor}_{\mathrm{C}}(-, x) \text { and } f \mapsto f_{*} \\
& H_{\mathrm{C}}^{\prime}: \mathrm{C} \longrightarrow \operatorname{Psh}\left(\mathrm{C}^{\mathrm{OP}}\right) \text {, mapping } x \mapsto \operatorname{Mor}_{\mathrm{C}}(x,-) \text { and } f \mapsto f^{*} .
\end{aligned}
$$

The functors $H_{\mathrm{C}}$ and $H_{\mathrm{C}}^{\prime}$ are called Yoneda functors.
Remark 1.7.3 ( $f_{*}$ and $f^{*}$ as natural transformations). The attentive reader may note that setting $H_{c} f=f_{*}$ is not quite right since $f_{*}$ is not a morphism in the category of presheaves $\operatorname{Psh}(\mathrm{C})$ according to our definition that roots from Lemma 1.3.2 - you are right, but we are being sloppy here just to simplify how we treat our objects and notations. The arrow $f_{*}$ (and $f^{*}$ ) is to be interpreted as the natural transformation $f_{*}: H_{C} x \Rightarrow H_{c} y$, where $f: x \rightarrow y$ - for which we define, for every object $z \in C$, the morphism $f_{*}: \operatorname{Mor}_{\mathrm{C}}(z, x) \rightarrow \operatorname{Mor}_{\mathrm{C}}(z, y)$ mapping $g \mapsto f g$.

Remark 1.7.4 (Size issues). Although C is a $\mathcal{U}$-category, its categories of presheaves described above need not be a $\mathcal{U}$-category - however, if C happens to be $\mathcal{U}$-small, then $\mathrm{Psh}(\mathrm{C})$ and $\mathrm{Psh}\left(\mathrm{C}^{\mathrm{op}}\right)$ are both $\mathcal{U}$-categories.

Lemma 1.7.5 (Yoneda). The Yoneda lemma states that:
(a) For all functors $F: \mathrm{C}^{\mathrm{op}} \rightarrow$ Set and $x \in \mathrm{C}$, there is a natural isomorphism:

$$
\operatorname{Mor}_{P s h(C)}\left(H_{C} x, F\right) \simeq F x
$$

(b) For all functors $G: C \rightarrow$ Set and $x \in \mathrm{C}$, there is a natural isomorphism:

$$
\operatorname{Mor}_{\mathrm{Psh}\left(\mathrm{C}^{\mathrm{op}}\right)}\left(G, H_{\mathrm{C}}^{\prime} x\right) \simeq G x
$$

These natural isomorphisms have a functorial nature - they define, respectively, functors $\mathrm{C}^{\mathrm{OP}} \times \mathrm{Psh}(\mathrm{C}) \rightarrow$ Set and $\mathrm{Psh}\left(\mathrm{C}^{\mathrm{OP}}\right)^{\mathrm{op}} \times \mathrm{C} \rightarrow$ Set

Proof. By means of ?? one only needs to prove one of the two statements, since the other follows from duality. We prove the item (a).

Let $\phi$ be the morphism making

commute. Notice that the choice of $\operatorname{id}_{x} \in \operatorname{Mor}_{\mathrm{C}}(x, x)=H_{\mathrm{C}} x(x)$ is done so that we can define a distinguished point $\eta_{x}\left(\mathrm{id}_{x}\right)$ from the set $F_{x}$.

Define now $\psi: F x \rightarrow \operatorname{Mor}_{P s h(C)}\left(H_{\mathrm{C}}(x), F\right)$ to be the map $F x \ni a \mapsto \eta^{a}: H_{\mathrm{C}} x \Rightarrow F$, where the natural transformation $\eta^{a}$ associated to $a$ is defined, for every object $y \in C$, by the map $\eta_{y}^{a}: \operatorname{Mor}_{C}(y, x) \rightarrow F x$ sending $f \mapsto F(f)(a)$.

Notice that any natural transformation $\eta \in \operatorname{Mor}_{\mathrm{Psh}(\mathrm{C})}\left(H_{\mathrm{C}} x, F\right)$ is completely determined by $\eta_{x}\left(\operatorname{id}_{x}\right)$. Indeed, given $f \in \operatorname{Mor}_{C}(y, x)$ one has the following diagram, which comes from the naturality of $\eta$ :


Thus we have $\eta_{y}(f)=(F f)\left(\eta_{x}\left(\mathrm{id}_{x}\right)\right)$ since $f^{*}\left(\mathrm{id}_{x}\right)=\mathrm{id}_{x} f=f$ — which shows that $\eta$ is determined by the distinguished map $\mathrm{id}_{x}$. That is, $\psi$ is determined, for each $a \in F x$ by $\eta_{x}^{a}\left(\mathrm{id}_{x}\right)$. This is thus sufficient to prove that $\phi$ and $\psi$ are mutual inverses and therefore stablish the wanted isomorphism.

Corollary 1.7.6. The Yoneda functors $H$ and $H^{\prime}$ are fully faithful.
Proof. From duality, we only check that the map $\operatorname{Mor}_{\mathrm{C}}(x, y) \rightarrow \operatorname{Mor}_{\mathrm{Psh}(\mathrm{C})}\left(H_{\mathrm{C}} x, H_{\mathrm{C}} y\right)$ sending $f \mapsto \eta^{f}$, where $\eta_{z}^{f}(g)=f_{*}(g)=f g$ for every $z \in \mathrm{C}$ and map $g \in \operatorname{Mor}_{\mathrm{C}}(z, x)$ is a bijection for all $x, y \in \mathrm{C}$. Such map is certainly injective, since given two distinct maps $f, h \in \operatorname{Mor}_{C}(x, y)$ we have $f_{*} \neq h_{*}$. Moreover, we have the isomorphism

$$
\operatorname{Mor}_{\mathrm{Psh}(\mathrm{c})}\left(H_{\mathrm{C}} x, H_{\mathrm{C}} y\right) \simeq H_{\mathrm{C}} y(x)=\operatorname{Mor}_{\mathrm{C}}(x, y)
$$

from Lemma 1.7.5. Let $\eta: H_{C} x \Rightarrow H_{C} y$ be any natural transformation - we wish to find $f \in \operatorname{Mor}_{\mathrm{C}}(x, y)$ for which $\eta_{z}(g)=f g$ for every $z \in \mathrm{C}$ and $g \in \operatorname{Mor}_{\mathrm{C}}(z, x)$. The natural choice is given by $f:=\eta_{x}\left(\mathrm{id}_{x}\right)$. By the naturality of $\eta$ we have, for any $w \in \mathrm{C}$


Thus given $h: w \rightarrow x$ in $C$ we have the equality

$$
\eta_{w}(h) g=\eta_{z}(h g)
$$

Therefore, if we restrict $w=x$ and $h=\mathrm{id}_{x}$ we get $\eta_{x}\left(\mathrm{id}_{x}\right) g=f g=\eta_{z}(g)$ - which is what we wanted, because this means that $f$ will have image $\eta$ under our mapping. \&

Remark 1.7.7 (Full subcategory of the presheaf category). From Corollary 1.7.6 one can conclude that C can be viewed as a full subcategory of the presheaf category $\operatorname{Psh}(\mathrm{C})$ (or of the category $\operatorname{Psh}\left(\mathrm{C}^{\mathrm{OP}}\right)$ ).

The following corollary establishes the main idea of the Yoneda lemma - one knows about an object by knowing how it interacts with other objects.

Corollary 1.7.8. Given a category $C$ and two objects $x, y \in C$. Then there exists an isomorphism $x \simeq y$ if and only if there exists an isomorphism $\operatorname{Mor}_{C}(-, x) \simeq \operatorname{Mor}_{C}(-, y)$ - or, for the covariant version, $\operatorname{Mor}_{\mathrm{C}}(x,-) \simeq \operatorname{Mor}_{\mathrm{C}}(y,-)$.

Proof. Since functors preserve isomorphisms, if $x \simeq y$ for any two objects $x, y \in C$, we obtain that $H_{\mathrm{C}} x \simeq H_{\mathrm{C}} y$. From Proposition 1.4.17 and since $H_{\mathrm{C}}$ is fully faithful we find that $H_{\mathrm{C}} x \simeq H_{\mathrm{C}} y$ implies $x \simeq y$.
Corollary 1.7.9. Let $F: \mathrm{C} \rightarrow \mathrm{D}$ be a functor of $\mathcal{U}$-categories and assume that C is $\mathcal{U}$-small. For every functor $A \in \operatorname{Psh}(C)$, the category $\mathrm{C}_{A}$ associated ${ }^{7}$ with the functor

$$
\mathrm{C} \xrightarrow{F} \mathrm{D} \xrightarrow{H_{\mathrm{D}}} \mathrm{Psh}(\mathrm{C})
$$

is $\mathcal{U}$-small. Analogously, given a functor $B \in \operatorname{Psh}\left(\mathrm{C}^{\mathrm{op}}\right)$, the category $\mathrm{C}^{B}$ associated with the functor

$$
\mathrm{C} \xrightarrow{F} \mathrm{D} \xrightarrow{H_{\mathrm{D}}^{\prime}} \operatorname{Psh}\left(\mathrm{C}^{\mathrm{op}}\right)
$$

is $\mathcal{U}$-small.
Proof. Recalling the definition, the category $C_{A}$ associated with $H_{C} F$ is given by the following collections:

$$
\begin{gathered}
\operatorname{Obj}\left(\mathrm{C}_{A}\right)=\left\{(x, \phi): x \in \mathrm{C} \text { and } \phi: H_{\mathrm{C}} F x \rightarrow A \text { in } \mathrm{D}\right\}, \\
\operatorname{Mor}_{\mathrm{C}_{A}}((x, \phi),(y, \psi))=\left\{f \in \operatorname{Mor}_{\mathrm{C}}(x, y): \phi=\psi H_{\mathrm{C}} F(f)\right\},
\end{gathered}
$$

where $(x, \phi),(y, \psi) \in C_{A}$ are any two objects, in other words, the collection of morphisms so that

commutes in D . Since C is assumed to be $\mathcal{U}$-small by hypothesis, it follows that the collection of arrows

$$
\coprod_{x \in \mathrm{C}} \operatorname{Mor}_{\mathrm{C}}(F x, A)
$$

is a disjoint union of $\mathcal{U}$-small sets, thus is itself $\mathcal{U}$-small. The same proof can be analogously constructed for the covariant case $C^{B}$.

Corollary 1.7.10. Let $C$ be a category and $f: x \rightarrow y$ a morphism in $C$. If for every $z \in C$ the morphism

$$
f_{*}: \operatorname{Mor}_{\mathrm{C}}(z, x) \rightarrow \operatorname{Mor}_{\mathrm{C}}(z, y)
$$

[^4]is an isomorphism ( or $f^{*}: \operatorname{Mor}_{\mathrm{C}}(x, z) \rightarrow \operatorname{Mor}_{\mathrm{C}}(y, z)$ for the covariant case), then $f$ is an isomorphism.

Proof. If the condition is met, then $H_{C} f: H_{C} x \rightarrow H_{C} y$ is an isomorphism. Since $H_{C}$ is fully faithful, then $f$ is an isomorphism. Notice that we already stated this proposition back when we were studying dual categories, with the Yoneda lemma we were able to prove it functorially - see Lemma 1.3.2.

## Functor Representation

Definition 1.7.11 (Representable functor). A presheaf $F: C^{o p} \rightarrow$ Set (or a functor $C \rightarrow$ Set) is said to be representable if there exists a natural isomorphism $H_{C} x \xlongequal{\simeq} F$ (or $\left.F \xlongequal{\Longrightarrow} H_{C}^{\prime} x\right)$ for some $x \in \mathrm{C}$ - the object $x$ is called a representative of $F$.

Corollary 1.7.12 (Representative uniqueness). The representative of a representable functor is unique up to unique isomorphism.

Proof. Since $H_{\mathrm{C}}$ (and $H_{\mathrm{C}}^{\prime}$ ) is fully faithful, it follows that, if $x, y \in \mathrm{C}$ are representatives of $F$, then there exists a natural isomorphism $H_{C} x \simeq H_{C} y$ hence, evoking Proposition 1.4.17, we find a unique isomorphism $x \simeq y$.

Corollary 1.7.13 (Universal element). Let $F: C^{\mathrm{op}} \rightarrow$ Set (or $C \rightarrow$ Set) be any representable presheaf, with representative object $x_{0} \in \mathrm{C}$. Then the natural isomorphism

$$
\eta: H_{c} x_{0} \xlongequal{\Longrightarrow} F
$$

is uniquely determined by an element $s_{0} \in F x_{0}$ - such an element is called a universal element of $F$.

Proof. Indeed, if $F$ is represented by $x_{0}$, then by the Yoneda lemma we obtain an isomorphism

$$
\operatorname{Mor}_{\operatorname{Psh}(\mathrm{C})}\left(H_{\mathrm{C}} x_{0}, F\right) \simeq F x_{0}
$$

therefore for each $s_{0} \in F x_{0}$ there exists a unique corresponding natural transformation $\eta: H_{\mathrm{C}} x_{0} \Rightarrow F$ (notice we didn't require it to be an isomorphism) and $\eta$ is completely determined by $s_{0}$ in the sense that, for every $y \in C$ and element $t \in F y$, there exists a unique morphism $f: x_{0} \rightarrow y$ such that $F(f)\left(s_{0}\right)=t$.

Corollary 1.7.14. Let $F: C \rightarrow P \operatorname{sh}(D)$ be a functor. If, for every $c \in C$, there exists an object $d \in \mathrm{D}$ such that $F c \simeq d$, then there exists a unique - up to unique isomorphism - functor $F_{0}: \mathrm{C} \rightarrow \mathrm{D}$ for which

$$
F \simeq H_{\mathrm{D}} F_{0} .
$$

Proof. From Remark 1.7.7 we know that D is a full subcategory of the presheaf category $\operatorname{Psh}(\mathrm{D})$. Since $H_{D}: D \rightarrow \operatorname{Psh}(\mathrm{D})$ is fully faithful, by means of Lemma 1.5.11 one finds
that there exists a unique functor $F_{0}: C \rightarrow D$ (up to unique isomorphism) and a natural isomorphism $F \simeq H_{\mathrm{D}} F_{0}$. In other words, the following diagram is quasi-commutative


## Properties of the Category of Elements

Proposition 1.7.15. Let $F: \mathrm{C}^{\mathrm{op}} \rightarrow$ Set be a functor (or $F: \mathrm{C} \rightarrow$ Set for the covariant case). Then $F$ is representable if and only if the category of elements $\mathrm{El}_{F}(\mathrm{C})$ has a final object (or initial object for the covariant case).

Proof. We prove the proposition for the contravariant case, the covariant case has a completely analogous construction and can be obtained by duality. Let $F$ be representable, with representative $x_{0} \in \mathrm{C}$, and universal element $s_{0}$ - that is, defining a natural isomorphism $H_{C} x_{0} \simeq F$. Consider any element $(x, s) \in \mathrm{El}_{F}(\mathrm{C})$ - the morphisms $(x, s) \rightarrow\left(x_{0}, s_{0}\right)$ in $\mathrm{El}_{F}(\mathrm{C})$ are arrows $f: x \rightarrow x_{0}$ in C for which $F(f)\left(s_{0}\right)=s$. Nonetheless, since $H_{\mathrm{C}}$ is fully faithful the morphism $f$ satisfying such condition is unique therefore $\left(x_{0}, s_{0}\right)$ is final in the category of elements $\mathrm{El}_{F}(\mathrm{C})$.

On the other hand, assume that $\mathrm{El}_{F}(\mathrm{C})$ has a final object $\left(x_{0}, s_{0}\right)$. We construct a natural transformation $\theta: F \Rightarrow H_{C} x_{0}$ by the maps $\theta_{x}: F x \rightarrow \operatorname{Mor}_{\mathrm{C}}\left(x, x_{0}\right)$, for $x \in \mathrm{C}$, sending $s \mapsto f$ where $F(f)\left(s_{0}\right)=s$ - this is well defined since $\left(x_{0}, s_{0}\right)$ is final and thus $f: x \rightarrow x_{0}$ is the unique morphism in $C$ with such property. Moreover, $\theta_{x}$ is clearly injective: if $s, s^{\prime} \in F x$ are any two elements such that $\theta_{x}(s)=\theta_{x}\left(s^{\prime}\right)=f$, then $F(f)\left(s_{0}\right)=s$ and also $F(f)\left(s_{0}\right)=s^{\prime}$ - which can only be the case for $s=s^{\prime}$. The surjectivity of $\theta_{x}$ is ensured by the fact that, given a morphism $f: x \rightarrow x_{0}$ in C , one can choose the element $F(f)\left(s_{0}\right) \in F x$ so that $\theta_{x}\left(F(f)\left(s_{0}\right)\right)=f$. We conclude that $\theta_{x}$ is a bijection and hence $\theta$ is a natural isomorphism $F \simeq H_{C} x_{0}$.

Definition 1.7.16 (Two sided represented functor). Let C be a $\mathcal{U}$-category, then there is a functor $\operatorname{Mor}(-,-): \mathrm{C}^{\mathrm{op}} \times \mathrm{C} \rightarrow$ Set defined by mapping objects $(x, y) \mapsto \operatorname{Mor}(x, y)$ and morphisms $(f, g):(a, y) \rightarrow(x, b)$ to a set-function $\left(f^{*}, g_{*}\right): \operatorname{Mor}(x, y) \rightarrow \operatorname{Mor}(a, b)$ given by the mapping $g \mapsto h g f$.

## Chapter 2

## Limits

### 2.1 Products

Definition 2.1.1 (Product). Let $J$ be a set, and $\left(C_{j}\right)_{j \in I}$ be a collection of objects in a category C.

A product of such collection, if it exists, is a pair $\left(P,\left(\pi_{j}\right)_{j \in J}\right)$-where $P$ is an object of C, and for every $j \in J, \pi_{j}: P \rightarrow C_{j}$ is a morphism of $C$.

Furthermore, this pair has to satisfy the following universal property: for every pair $\left(Q,\left(q_{j}\right)_{j \in J}\right)$ —where $Q \in C$ and for every $j \in J, q_{j}: Q \rightarrow C_{j}$ is a morphism in C-there exists a unique morphism $f: Q \rightarrow P$ of $C$ such that, for every $j \in J$ the following diagram commutes


Proposition 2.1.2 (Uniqueness). The product of a collection of objects, if existent, is unique up to isomorphism.

Proof. Let C be a category admitting the product of a collection $\left(C_{j}\right)_{j \in J}$ of objects of C , and $J$ is a set. Let $\left(P,\left(p_{j}\right)_{j \in J}\right)$ and $\left(Q,\left(q_{j}\right)_{j \in J}\right)$ be products of $\left(C_{j}\right)_{j \in J}$ in the category C . Since $P$ and $Q$ are products, there exists unique morphisms $f: Q \rightarrow P$ and $g: P \rightarrow Q$ such that, for all $j \in J$, we have $q_{j}=p_{j} f$ and $p_{j}=q_{j} g$.

Moreover, one can apply the product property of $P$ to $P$ itself: there exists a unique morphism $h: P \rightarrow P$ such that, for all $j \in J$, we have $p_{j}=p_{j} h$. For that to be true, it is clear that we must have $h=\operatorname{id}_{P}$. Note, however, that for each $j \in J$,

$$
p_{j}=q_{j} g=p_{j} f g .
$$

Since $h$ is unique, we obtain $f g=h=\operatorname{id}_{p}$. On the other hand, applying the product property of $Q$ in $Q$ will yield $g f=\operatorname{id}_{Q}$. This shows that $Q \simeq P$ in C , via $f$ and $g$.

Proposition 2.1.3 (Products are independent of ordering). Let $I$ be a set and $\left(J_{k}\right)_{k \in K}$ be a partition of $I$ by disjoint subsets $J_{k} \subseteq I$. Let $\left(C_{i}\right)_{i \in I}$ be a collection of objects in a
category C. Then, if all the products presented below exist in C, they are isomorphic:

$$
\prod_{i \in I} C_{i} \simeq \prod_{k \in K}\left(\prod_{j \in J_{k}} C_{j}\right)
$$

Proof. Define the following collections:

- Let $\left(q_{k}\right)_{k \in K}$ be the collection of morphisms $q_{k_{0}}: \prod_{k \in K}\left(\prod_{j \in J_{k}} C_{j}\right) \rightarrow C_{k_{0}}$, for each $k_{0} \in K$, associated with the product $\prod_{k \in K}\left(\prod_{j \in J_{k}} C_{j}\right)$ in C .
- For all $k \in K$, let $\left(p_{j}\right)_{j \in J_{k}}$ be the collection of morphisms $p_{j_{0}}: \prod_{j \in J_{k}} C_{j} \rightarrow C_{j_{0}}$, for each $j_{0} \in J_{k}$ associated with the product $\prod_{j \in J_{k}} C_{j}$ in $C$.

Let $\left(L,(\ell)_{i \in I}\right)$ be a pair where $L \in C$ and $\ell_{i}: L \rightarrow C_{i}$ is a morphism in $C$ for each $i \in I$.
Define a collection $\left(g_{k}\right)_{k \in K}$ where, for each $k_{0} \in K$, we let $g_{k_{0}}: L \rightarrow \prod_{j \in J_{k_{0}}} C_{j}$ be the unique morphism of C such that $\ell_{i}=g_{k_{0}} p_{i}$ for every $i \in J_{k}$. Let $f: \prod_{k \in K}\left(\prod_{j \in J_{k}} C_{j}\right)$ be the unique morphism of $C$ such that $g_{k}=q_{k} f$.

We see that the following diagram commutes for every $k_{0} \in K$ and $i \in J_{k}$ :


Since $I=\bigcup_{k \in K} J_{k}$, the diagram commutes for any $i \in I$-thus $\prod_{k \in K}\left(\prod_{j \in J_{k}} C_{j}\right)$ is a product of the collection $\left(C_{i}\right)_{i \in I}$. Now, Proposition 2.1.2 finishes the proof.

### 2.2 Coproducts

Definition 2.2.1 (Coproduct). Let $J$ be a set, and $\left(C_{j}\right)_{j \in I}$ be a collection of objects in a category C.

A coproduct of such collection, if it exists, is a pair $\left(P,\left(\iota_{j}\right)_{j \in J}\right)$-where $P$ is an object of $C$, and for every $j \in J, \iota j: P \rightarrow C_{j}$ is a morphism of $C$.

Furthermore, this pair has to satisfy the following universal property: for every pair $\left(Q,\left(q_{j}\right)_{j \in J}\right)$ —where $Q \in C$ and for every $j \in J, q_{j}: C_{j} \rightarrow Q$ is a morphism in C-there exists a unique morphism $f: P \rightarrow Q$ of $C$ such that, for every $j \in J$ the following diagram
commutes


We can see right away that coproducts are dual to products, yielding the following two dual properties.

Proposition 2.2.2 (Uniqueness). If a coproduct of a collection of objects of a category exists, then this coproduct is unique up to isomorphism.

Proposition 2.2.3 (Coproducts are independent of ordering). Let $I$ be a set and $\left(J_{k}\right)_{k \in K}$ be a partition of $I$ by disjoint subsets $J_{k} \subseteq I$. Let $\left(C_{i}\right)_{i \in I}$ be a collection of objects in a category C . Then, if all the coproducts presented below exist in C , they are isomorphic:

$$
\coprod_{i \in I} C_{i} \simeq \coprod_{k \in K}\left(\coprod_{j \in J_{k}} C_{j}\right)
$$

Remark 2.2.4 (Duality). Coproducts are dual to products in the following sense: if $P$ is a product in a category $C$, then $P^{\text {op }}$ is a coproduct in the opposite category C ${ }^{\text {op }}$. Notice however that one cannot simply reverses the arrows of a product in C and end up with a coproduct in C—in fact, there are categories where product do exist, while coproducts do not.

### 2.3 Equalizers and Coequalizers

Definition 2.3.1 (Equalizer). Let $f, g: A \rightrightarrows B$ be parallel morphisms in a category $C$. An equalizer of $f$ and $g$ is a pair $(K, k)$-where $K \in C$ is an object and $k: K \rightarrow A$ is a morphism of $C$ for which $f k=g k$-such that, for any object $M \in C$ and morphism $m: M \rightarrow A$ of $C$ satisfying $f m=g m$, there exists a unique morphism $n: M \rightarrow K$ such that the following diagram commutes


Just as with products and coproducts, one has a dual notion of an equalizer, we call it a coequalizer. For the sake of later reference, we write down its definition.

Definition 2.3.2 (Coequalizer). Let $f, g: A \rightrightarrows B$ be parallel morphisms in a category C. An equalizer of $f$ and $g$ is a pair $(C, c)$-where $C \in C$ is an object and $c: B \rightarrow C$ is a morphism of $C$ for which $c f=c g-s u c h$ that, for any object $M \in \mathrm{C}$ and morphism
$m: B \rightarrow M$ of $C$ satisfying $m f=m g$, there exists a unique morphism $n: C \rightarrow M$ such that the following diagram commutes


Proposition 2.3.3 (Uniqueness). Given two parallel morphisms $f, g: A \rightrightarrows B$ in a category C, if the (co)equalizer of them exists, then it is unique up to isomorphism. We denote the equalizer of $f$ and $g$ by eq $(f, g)$ and the coequalizer of $f$ and $g$ by coeq $(f, g)$.

Proof. We prove for equalizers. Let $(K, k)$ and $\left(K^{\prime}, k^{\prime}\right)$ be equalizers of $f$ and $g$. If we apply the equalizer property of $\left(K^{\prime}, k^{\prime}\right)$ in $(K, k)$ and vice versa, we obtain unique morphisms $n^{\prime}: K \rightarrow K^{\prime}$ and $n: K^{\prime} \rightarrow K$ such that $k=k^{\prime} n^{\prime}$ and $k^{\prime}=k n$.

Moreover, applying the equalizer property of $(K, k)$ in itself we obtain a unique $t: K \rightarrow K$ such that $k=k t$-therefore $t=\operatorname{id}_{K}$. Since $k=k^{\prime} n=k \mathrm{id}_{K}$ and $\mathrm{id}_{K}$ is unique with such property, it follows that, since $k=k^{\prime} n^{\prime}=(k n) n^{\prime}=k\left(n n^{\prime}\right)$ we find that $n n^{\prime}=\operatorname{id}_{K}$. On the other hand, doing the same for $\left(K^{\prime}, k^{\prime}\right)$ implies in $n^{\prime} n=\operatorname{id}_{K^{\prime}}$. Therefore $K \simeq K^{\prime}$ in C , via $n$ and $n^{\prime}$.

Corollary 2.3.4. Given a morphism $f: A \rightarrow B$ in a category $C$, the (co)equalizer of $f$ with itself always exists-in fact eq $(f, f)=\left(A, \operatorname{id}_{A}\right)$ and $\operatorname{coeq}(f, f)=\left(B, \mathrm{id}_{B}\right)$.

Proposition 2.3.5. Let $C$ be a category and $f, g: A \rightrightarrows B$ be parallel morphisms in $C$. The following are properties concerning equalizers and coequalizers:

1. If $f$ and $g$ have an equalizer eq $(f, g)$ in C, the morphism eq $(f, g) \mapsto A$ associated to the equalizer is a monomorphism.
2. If $f$ and $g$ have an coequalizer coeq $(f, g)$ in $C$, the morphism $B \rightarrow \operatorname{coeq}(f, g)$ associated to the coequalizer is an epimorphism.

Proof. The properties are dual to each other, thus we may simply prove the one concerning equalizers. Let $k: \mathrm{eq}(f, g) \rightarrow A$ be the morphism in $C$ associated to the equalizer. Consider any two parallel morphisms $x, y: C \rightrightarrows \mathrm{eq}(f, g)$ such that $k x=k y$-therefore, the following diagram commutes

$$
C \xrightarrow[y]{\stackrel{x}{\longrightarrow}} \mathrm{eq}(f, g) \xrightarrow{k} A \xrightarrow[g]{\stackrel{f}{\longrightarrow}} B
$$

From the diagram we obtain the relation $f(k x)=g(k x)$. Since $k x: C \rightarrow A$, the universal property implies that $x$ must be the unique-hence $y=x$.

Proposition 2.3.6. Let $f: A \rightarrow B$ be a morphism of a category C. If $f$ is both an epimorphism and equalizer ${ }^{1}$, then $f$ is an isomorphism.

[^5]Proof. Let $x, y: B \rightrightarrows C$ be parallel morphisms in $C$ with eq $(x, y)=f$, then

$$
A \xrightarrow{f} B \xrightarrow[g]{\stackrel{f}{\Longrightarrow}} C
$$

which implies in $x f=y f$. By hypothesis $f$ is an epimorphism, thus $x=y$. Notice however that the equalizer of a morphism $B \rightarrow C$ with itself is the identity on $B$. Since equalizers are unique up to isomorphism, it follows that $f \simeq \operatorname{id}_{B}$ —thus $f$ itself is an isomorphism (and $A \simeq B$ ).

### 2.4 Pullbacks \& Pushouts

Definition 2.4.1 (Pullback). Let $f: A \rightarrow C$ and $g: B \rightarrow C$ be any two morphisms in a category C. A pullback of $(f, g)$ is a triple $\left(P, f^{\prime}, g^{\prime}\right)$ where the diagram

commutes in C, and for every other triple ( $Q, f^{\prime \prime}, g^{\prime \prime}$ ) making the diagram

commute in C , there exists a unique morphism $\ell: Q \rightarrow P$ in C such that the following diagram commutes


We usually denote that a square is a pullback in C by marking it as follows


The dual notion of a pullback is that of a pushout, just as before, we'll write it down just for the sake of later reference.

Definition 2.4.2 (Pushout). Let $f: C \rightarrow A$ and $g: C \rightarrow B$ be any two morphisms in a category C. A pushout of $(f, g)$ is a triple $\left(P, f^{\prime}, g^{\prime}\right)$ where the diagram

commutes in C, and for every other triple $\left(Q, f^{\prime \prime}, g^{\prime \prime}\right)$ making the diagram

commute in C , there exists a unique morphism $\ell: P \rightarrow Q$ in C such that the following diagram commutes


Analogously, if we want to visually say that a square is a pushout in C, we mark it as follows


Proposition 2.4.3 (Uniqueness). The pullback (or pushout) of two morphisms, if existent, is unique up to isomorphism.

Proof. Let $f: A \rightarrow C$ and $g: B \rightarrow C$ be two morphisms in a category C. Suppose there exists two pullbacks of $(f, g)$, namely $\left(P, f^{\prime}, g^{\prime \prime}\right)$ and $\left(Q, f^{\prime \prime}, g^{\prime \prime}\right)$. Let $\phi: P \rightarrow Q$ and $\psi: Q \rightarrow P$ be the uniquely defined morphisms given by the universal property of the
pullback. Consider the following two commutative diagrams


Notice, however, that since $\operatorname{id}_{P}: P \rightarrow P$ and $\operatorname{id}_{Q}: Q \rightarrow Q$ also make them commuterespectively, the left and right diagrams. From uniqueness we obtain $\phi \psi=\mathrm{id}_{P}$ and $\psi \phi=\operatorname{id}_{Q}$-therefore $P \simeq Q$ in C , via $\phi$ and $\psi$.

Proposition 2.4.4. Let $\left(P, f^{\prime}, g^{\prime}\right)$ be the pullback of a pair of morphisms $(f, g)$ in a category C, then:
(a) If $g$ is a monomorphism, then $g^{\prime}$ is also a monomorphism.
(b) If $g$ is an isomorphism, then $g^{\prime}$ is also an isomorphism.

Dually, if $\left(B, f^{\prime \prime}, g^{\prime \prime}\right)$ is the pushout of the pair $(f, g)$, then:
(c) If $g$ is an epimorphism, then $g^{\prime \prime}$ is also an epimorphism.
(d) If $g$ is an isomorphism, then $g^{\prime \prime}$ is also an isomorphism.

Proof. We only prove items (a) and (b), since (c) and (d) are merely dual consequences of the former items.
(a) Let $g$ be a monomorphism and consider parallel morphisms $u, v: Q \rightrightarrows P$ such that $g^{\prime} u=g^{\prime} v$-we want to prove that $u=v$. Take into account the following commutative diagram


Lets first consider the morphism $u$. Notice that $p:=g^{\prime} u$ and $q:=f^{\prime} u$ are such that $f p=g q$. From the universal property of the pullback we have that $u$ is the unique morphism factorizing $(p, q)$ through $(g, f)$.

On the other hand, if we consider the arrow $v$, one can define $p^{\prime}:=g^{\prime} v$ and $q^{\prime}:=f^{\prime} v$ so that $f p^{\prime}=g q^{\prime}$-thus $v$ is the unique factorization of $\left(p^{\prime}, q^{\prime}\right)$ through $(g, f)$. Notice, however, that from construction $p^{\prime}=g^{\prime} v=g^{\prime} u=p$. Moreover,

$$
g q^{\prime}=f p^{\prime}=f\left(g^{\prime} v\right)=f\left(g^{\prime} u\right)=f p=g q
$$

since $g$ is a monomorphism, it follows that $g q^{\prime}=g q$ implies $q^{\prime}=q$. Therefore both $u$ and $v$ are factorizations of the same pair of morphisms-and from uniqueness, it can only be the case that $u=v$.
(b) Suppose $g$ is an isomorphism and consider the following commutative diagram


Thus $g^{\prime} \ell=\mathrm{id}_{A}$ is already given. We can now consider the diagram


Moreover, we consider the following composition

$$
g^{\prime}\left(\ell g^{\prime}\right)=\left(g^{\prime} \ell\right) g^{\prime}=\operatorname{id}_{A} g^{\prime}=g^{\prime}=g^{\prime} \operatorname{id}_{P}
$$

on the other hand we have

$$
f^{\prime}\left(\ell g^{\prime}\right)=\left(f^{\prime} \ell\right) g^{\prime}=\left(g^{-1} f\right) g^{\prime}=g^{-1}\left(f g^{\prime}\right)=g^{-1}\left(g f^{\prime}\right)=f^{\prime}=f^{\prime} \operatorname{id}_{p}
$$

Therefore $g^{\prime}\left(\ell g^{\prime}\right)=g^{\prime} \mathrm{id}_{P}$ and $f^{\prime}\left(\ell g^{\prime}\right)=f^{\prime} \mathrm{id}_{P}$ but by uniqueness, since $\mathrm{id}_{P}$ also makes Eq. (2.1) commute, it follows that $\ell g^{\prime}=\operatorname{id}_{P}$. Thus $g^{\prime}$ is an isomorphism with inverse $\ell$.

Definition 2.4.5 (Kernel \& cokernel). Let $f: A \rightarrow B$ be a morphism in a category C. The (co)kernel of $f$, if existent, is defined to be the pullback of $f$ with itself-or pushout in the case of cokernels.

Proposition 2.4.6. Let $C$ be a category and $f: A \rightarrow B$ be a morphism of $C$. If the (co)kernel of $f$ exists, its associated morphisms are both epimorphisms (monomorphisms for the case of cokernels).

Proof. Let ker $f:=(K, \alpha, \beta)$ and consider $A$ itself, together with the identity morphisms. Since $K$ is a pullback of $f$ with itself, there exists a unique morphism $\gamma: A \rightarrow K$ such that $\beta \gamma=\mathrm{id}_{A}=\alpha \gamma$-therefore $\gamma$ is a split monomorphism and both $\alpha$ and $\beta$ are split epimorphisms.

Proposition 2.4.7. Let $f: A \rightarrow B$ be a morphism in a category $C$. Then the following properties are equivalent:
(a) The morphism $f$ is monic (conversely, epic).
(b) The kernel of $f$ exists, furthermore $\operatorname{ker} f=\left(A, \mathrm{id}_{A}, \mathrm{id}_{A}\right)$ (conversely we have coker $\left.f=\left(B, \mathrm{id}_{B}, \mathrm{id}_{B}\right)\right)$.
(c) The (co)kernel $(K, \alpha, \beta)$ of $f$ exists, and $\alpha=\beta$.

Proof. - (a) $\Rightarrow(\mathrm{b})$ : Since $f$ is monic, if $P$ is any object together with morphisms $\phi, \psi: P \rightrightarrows A$ such that $f \phi=f \psi$ then $\phi=\psi$ and we may simply take $A$ to be the kernel of $f$ together with the identity morphisms-while the unique morphism $P \rightarrow A$ is given by $\phi=\psi$.

- $(\mathrm{b}) \Rightarrow(\mathrm{c})$ : If $(K, \alpha, \beta)$ is a kernel for $f$ then in particular there exists an isomorphism $\phi: A \rightarrow K$ and therefore $\alpha \phi=\operatorname{id}_{A}$ and $\beta \phi=\operatorname{id}_{A}$. Since $\phi$ is epic, it follows that $\alpha=\beta$.
- $(\mathrm{c}) \Rightarrow(\mathrm{a})$ : Suppose $(K, \alpha, \alpha)$ is kernel for $f$, then for all objects $P$ and morphisms $g, h: P \rightrightarrows A$ such that $f g=f h$, we have a unique $\phi: P \rightarrow K$ such that $\alpha \phi=g$ and $\alpha \phi=h$-thus $g=h$.

Proposition 2.4.8 (Coequalizers \& kernels). Let C be a category. The following properties relate coequalizers and kernels:
(a) Consider parallel morphisms $x, y: X \rightrightarrows A$. If the coequalizer $f=\operatorname{coeq}(x, y)$ exists and has a kernel pair $\operatorname{ker} f=(\alpha, \beta)$, then $\operatorname{coeq}(x, y)$ is the coequalizer of $\operatorname{ker} f$.
(b) Consider a morphism $h: A \rightarrow B$. If the kernel ker $h=(\varepsilon, \delta)$ exists and has a coequalizer $w=\operatorname{coeq}(\varepsilon, \delta)$, then $\operatorname{ker} h$ is the kernel of coeq $(\varepsilon, \delta)$.

Proof. (a) Since $f$ is the coequalizer of $x$ and $y$, it cleartly satisfies $f x=f y$. Therefore, the triple $(X, x, y)$ can be used to apply the universal property of the pullback of $\operatorname{ker} f$ to obtain a unique factorization $\phi: X \rightarrow \operatorname{ker} f$ such that $x=\alpha \phi$ and $y=\beta \phi$.
Notice that $\alpha, \beta$ : $\operatorname{ker} f \rightrightarrows A$ are parallel morphisms, so one can be tempted to find its coequalizer-we shall prove that $\operatorname{coeq}(\alpha, \beta)=\operatorname{coeq}(x, y)$. Let $C$ be an object together with a morphism $g: A \rightarrow C$ such that $g \alpha=g \beta$. We may precompose this morphism with $\phi$, obtaining $g \alpha \phi=g \beta \phi$, but using the result for the last paragraph we conclude that $g x=g y$. Using the universal property for the coequalizer of
$x$ and $y$, we find a unique factorization morphism $\psi: \operatorname{coeq}(x, y) \rightarrow C$ such that $g=\psi f$. Therefore, $\operatorname{coeq}(x, y)$ satisfies the universal property for the coequalizer of the kernel pair $(\alpha, \beta)$ of $f$. The whole construction of this item's proof can be seen in the following commutative diagram:

(b) From the pullback definition, we know that $h \varepsilon=h \delta$-since $w$ is the coequalizer of $(\varepsilon, \delta)$, it follows that there exists a unique morphism $\gamma: B \rightarrow \operatorname{coeq}(\varepsilon, \delta)$ such that $\gamma w=h$.
Consider now two parallel morphisms $a, b: Y \rightrightarrows A$ such that $w a=w b$. Then we have that $h a=(\gamma w) a$, and $h b=(\gamma w) b$ since $w b=w a$ by assumption, it follows that $h a=h b$. From the kernel property, there must exist a unique morphism $\eta: Y \rightarrow$ $\operatorname{ker} h$ such that $\varepsilon \eta=a$ and $\delta \eta=b$. Therefore $\operatorname{ker} w=\operatorname{ker} h$. All constructions can be visualized in the following commutative diagram:


Proposition 2.4.9 (Associativity property). Let C be a category and consider the following commutative diagram in C:


The following are properties which regard pullbacks of such commutative squares:
(a) If both squares are pullbacks, then the outer-square

is a pullback.
(b) If the second square is a pullback and the outer diagram Eq. (2.2) is a pullback, then the first square is a pullback.

Proof. (a) Let $x: Z \rightarrow D$ and $y: Z \rightarrow C$ be two morphisms of $C$ such that $g f x=e y$. From the pullback property of the second square, there exists a unique morphism $z: Z \rightarrow B$ such that $b z=y$ and $d z=f x$. Using the pullback property of the first square on the morphisms $z$ and $x$ we find a unique $w: Z \rightarrow A$ such that $a w=z$ and $c w=x$.

Notice that since $a w=z$ then $b(a w)=b z=y$-we'll show that $w$ is the unique morphism of C such that $(b a) w=y$ and $c w=x$. Suppose the existence of another morphism $w^{\prime}: Z \rightarrow A$ such that $(b a) w^{\prime}=y$ and $c w^{\prime}=x$. In particular, it follows that $b\left(a w^{\prime}\right)=b(a w)$, and

$$
d\left(a w^{\prime}\right)=f\left(c w^{\prime}\right)=f x=f(c w)=d(a w)
$$

By the uniqueness of the morphism $Z \rightarrow B$ from the second pullback square diagram, $a w^{\prime}=z=a w$. Using this last equality and the fact that $c w^{\prime}=x=c w$, by the pullback property of the first square we obtain $w^{\prime}=w$-which finally settles that $w$ is unique and the outer-square is indeed a pullback.
(b) We assume the existence of a pullback $\left(A^{\prime}, c^{\prime}, a^{\prime}\right)$ of $(f, d)$ and show that $A^{\prime}$ must be isomorphic to $A$. From the commutativity of the diagram, since $d a=c f$, we use the pullback property of $\left(A^{\prime}, c^{\prime}, a^{\prime}\right)$ to get a unique morphism $h: A \rightarrow A^{\prime}$-we'll show that $h$ is the required isomorphism. Via the result of item (a), we know that $\left(A^{\prime}, c, b a^{\prime}\right)$ is a pullback for the outer-square, therefore it follows that $c^{\prime} h=c$ and $b a^{\prime} h=b a$. Hence from the last two equalities one concludes that $h$ is a factorization between two pullbacks of the outer-square. Since $h$ is unique and pullbacks are unique up to isomorphism, it must be the case that $h$ is an isomorphism-thus $A \simeq A^{\prime}$, then $(A, c, a)$ is a pullback for the first square.

### 2.5 Limits \& Colimits

## (Co)Cones \& (Co)Limits

Definition 2.5.1 (Cone). Let $F: \mathrm{D} \rightarrow \mathrm{C}$ be a functor. We define a cone on $F$ to consist of the following data:

- An object $C \in C$.
- For each object $D \in \mathrm{D}$, a corresponding morphism $p_{D}: C \rightarrow F D$ in $C$. Moreover, for every morphism $d: D \rightarrow D^{\prime}$ in D , one has that $p_{D^{\prime}}=F d \circ p_{D}$.

Definition 2.5.2 (Limit of a functor). Let $F: \mathrm{D} \rightarrow \mathrm{C}$ be a functor. We define a limit of $F$ to be a cone $\lim F=\left(L,\left(p_{D}\right)_{D \in \mathrm{D}}\right)$ with the property that, for any cone $\left(M,\left(q_{D}\right)_{D \in \mathrm{D}}\right)$ on $F$, there exists a unique morphism $m: M \rightarrow L$ of $C$ such that, for every $D \in D$, the following diagram commutes


Proposition 2.5.3 (Uniqueness). The limit of a functor, when existent, is unique up to isomorphism.

Proof. Let $F: \mathrm{D} \rightarrow \mathrm{C}$ be a functor admitting limit cones $\left(L,\left(p_{D}\right)_{D \in \mathrm{D}}\right)$ and $\left(L^{\prime},\left(p_{D}^{\prime}\right)_{D \in \mathrm{D}}\right)$. Then there exists unique morphisms $m: L^{\prime} \rightarrow L$ and $m: L \rightarrow L^{\prime}$ such that $p_{D} m=p_{D}^{\prime}$ and $m^{\prime} p_{D}^{\prime}=p_{D}$. Stacking these two we find that $m m^{\prime}: L \rightarrow L$ is such that $p_{D}\left(m m^{\prime}\right)=p_{D}$, but since $\mathrm{id}_{L}$ has the same property, by the uniqueness of the factoring morphism, it follows that $m m^{\prime}=\mathrm{id}_{L}$. The same analogous argument goes for $m^{\prime} m=\mathrm{id}_{L^{\prime}}$. We conclude that $L \simeq L^{\prime}$ in $C$ via $m$ and $m^{\prime}$.

Proposition 2.5.4 (Parallel factorizations). Let $F: \mathrm{D} \rightarrow \mathrm{C}$ be a functor admitting a limit, and let $M$ be any object in C. Two parallel morphisms $f, g: M \rightrightarrows \lim F$ in C are equal if, for every $D \in \mathrm{D}$, one has $p_{D} f=p_{D} g$.

Proof. Notice that $\left(M,\left(p_{D} f\right)_{D \in \mathrm{D}}\right)$ forms a cone on $F$, therefore, from the universal property of the limit of $F$, we obtain $f=g$.

Definition 2.5.5 (Cocone). Let $F: \mathrm{D} \rightarrow \mathrm{C}$ be a functor. We define a cocone on $F$ to consist of the following data:

- An object $C \in C$.
- For each object $D \in D$, a corresponding morphism $s_{C}: F D \rightarrow C$ of $C$ such that, for every morphism $d: D^{\prime} \rightarrow D$ in $D$, we have that $s_{D^{\prime}}=s_{D} \circ F d$.

Definition 2.5.6 (Colimit). Let $F: \mathrm{D} \rightarrow \mathrm{C}$ be a functor. We define a colimit on $F$ to be a cocone colim $F=\left(L,\left(s_{D}\right)_{D \in \mathrm{D}}\right)$ such that, for every cocone $\left(M,\left(t_{D}\right)_{D \in \mathrm{D}}\right)$ on $F$, there exists a unique morphism $m: L \rightarrow M$ such that, for every $D \in D$, the following diagram commutes


Example 2.5.7 (Products). Consider a set $I$ as a discrete category (as in Example 1.2.10), and any category $C$. The limit of a functor $F: I \rightarrow C$, if existent, is simply a product in C , that is

$$
\lim F \simeq \prod_{i \in I} F i
$$

Example 2.5.8 (Inverse \& direct limit). The limit of a functor $F: \omega^{\mathrm{op}} \rightarrow \mathrm{C}$ is commonly referred to as an inverse limit:


The dual of this concept is that of the direct limit, which is the colimit of a functor $G: \omega \rightarrow C$, visually given by:


## Complete Categories

Proposition 2.5.9 (All limits). Let C be a $\mathcal{U}$-category. If for all $\mathcal{U}$-categories D and functors $F: D \rightarrow C$, the limit of $F$ exists in $C$-that is, the category admits all limits-then $C$ is a preorder class.

Proof. For C to be a preorder, there must exist at most one morphism between every pair of objects of C , so this is what we settle to do. Let $A, B \in \mathrm{C}$ be any two objects and suppose there exists a pair of distinct parallel morphisms $f, g: A \rightrightarrows B$. Since every limit exists, then in particular the product $B^{|\operatorname{Mor}(\mathrm{C})|}$ is a well defined object of $C$. From $f$ and $g$, one can create $2^{|\operatorname{Mor}(\mathrm{C})|}$ distinct collections of morphisms $(h: A \rightarrow B)|\operatorname{Mor}(\mathrm{C})|$ —where each $h$ is either $f$ or $g$-which are cones over an "inclusion" functor $F: \mathrm{A} \rightarrow \mathrm{C}$, where A is composed of objects $A$ and $B$, and morphisms are $f, g: A \rightrightarrows B$. The admittance of a limit over this functor is to state the existence of $2^{|\operatorname{Mor}(\mathrm{C})|}$ distinct factorizations of $A \rightarrow B^{|\operatorname{Mor}(C)|}$ in C . Since all of these factorizations are morphisms of C , it should be the case that $2^{|\operatorname{Mor}(\mathrm{C})|}<|\operatorname{Mor}(\mathrm{C})|$, which contradicts Cantor's theorem (see Theorem B.1.1) since C is a $\mathcal{U}$-category. If follows that there cannot exist more than one morphism between the objects of $C$, proving that it is a preorder.

Definition 2.5.10 (Completeness). We define the following notions concerning categories and the existence of limits:
(a) A category C is said to be (co)complete if, for every small category D, any functor $F: D \rightarrow C$ has a (co)limit in C.
(b) A category C is said to be finitely (co)complete if, for every finite category D , any functor $F: \mathrm{D} \rightarrow \mathrm{C}$ has a (co)limit in C .

## Existence Theorem for Limits

Theorem 2.5.11. A category C is complete if and only if each collection of objects, indexed by a set, has a product and each pair of parallel morphisms has an equalizer.

Proof. Let C be complete and $I$ be any set, which defines a small category. If $\left(C_{i}\right)_{i \in I}$ is any collection of objects in $C$, one can define a functor $F: I \rightarrow C$ by $F i:=C_{i}$ and, since $F$ has a limit by hypothesis, it follows that $\prod_{i \in I} F i=\prod_{i \in I} C_{i}$ is a product in C. For the equalizer, let $f, g: A \rightrightarrows B$ be two parallel morphisms in C. Define a category D whose objects are $A$ and $B$, and morphisms are the identities together with both $f$ and $g$. Defining $F: \mathrm{D} \rightarrow \mathrm{C}$ to be simply an inclusion, since $F$ has a limit, then there exists $\lim F$ such that

$$
\lim F \longrightarrow A \underset{g}{\stackrel{f}{\Longrightarrow}} B
$$

is an equalizer in C , which proves the first proposition.
For the second proposition, suppose that $C$ admits products and equalizers. Let $D$ be a small category and $F: \mathrm{D} \rightarrow \mathrm{C}$ be any functor. Define the following:

- Consider two pairs of products

$$
\left(\prod_{D \in \mathrm{D}} F D,\left(r_{D}\right)_{D \in \mathrm{D}}\right) \quad \text { and } \quad\left(\prod_{f \in \operatorname{Mor}(\mathrm{D})} F(\operatorname{cod} f),\left(q_{f}\right)_{f \in \operatorname{Mor}(\mathrm{D})}\right),
$$

where $r_{D^{\prime}}: \prod_{D \in \mathrm{D}} F D \rightarrow F D^{\prime}$ and $q_{f^{\prime}}: \prod_{f \in \operatorname{Mor}(\mathrm{D})} F(\operatorname{cod} f) \rightarrow F\left(\operatorname{cod} f^{\prime}\right)$ are the projections associated with the products.

- Let $\alpha, \beta: \prod_{D \in \mathrm{D}} F D \rightrightarrows \prod_{f \in \operatorname{Mor}(\mathrm{D})} F(\operatorname{cod} f)$ be the unique factorizations such that

$$
q_{f} \alpha=r_{\operatorname{cod} f} \quad \text { and } \quad q_{f} \beta=F f \circ r_{\operatorname{dom} f}
$$

for all $f \in \operatorname{Mor}(\mathrm{D})$.

- Since equalizers always exist on C , let $(L, \ell):=\mathrm{eq}(\alpha, \beta)$. Define a collection $\left(p_{D}\right)_{D \in \mathrm{D}}$ for which $p_{D}:=r_{D} \ell$. We shall prove that $\lim F=\left(L,\left(p_{D}\right)_{D \in D}\right)$.

We first prove that $\left(L,\left(p_{D}\right)_{D \in D}\right)$ is a cone on $F$. To that end, consider any morphism $f: D \rightarrow D^{\prime}$ in $D$. From the construction of the collection of morphisms, we have that

$$
\begin{aligned}
F f \circ p_{D} & =F f \circ\left(r_{D} \ell\right)=\left(F f \circ r_{D}\right) \ell \\
& =\left(q_{f} \beta\right) \ell=q_{f}(\beta \ell) \\
& =q_{f}(\alpha \ell)=\left(q_{f} \alpha\right) \ell \\
& =r_{D^{\prime}} \ell \\
& =p_{D^{\prime}}
\end{aligned}
$$

therefore, $\left(p_{D}\right)_{D \in \mathrm{D}}$ satisfies the conditions of a cone on $F$.
Let $\left(M,\left(h_{D}\right)_{D \in \mathrm{D}}\right)$ be a cone on $F$. From the product universal property, there exists a unique factorization $\gamma: M \rightarrow \prod_{D \in D} F D$ such that $r_{D} \gamma=h_{D}$. On the other hand, for any morphism $f: D \rightarrow D^{\prime}$ of $D$ we have

$$
\left(q_{f} \alpha\right) \gamma=r_{D^{\prime}} \gamma=h_{D^{\prime}}=F f \circ h_{D}=F f \circ\left(r_{D} \gamma\right)=\left(F f \circ r_{D}\right) \gamma=\left(q_{f} \beta\right) \gamma
$$

Since the factorization on the product $\prod_{f \in \operatorname{Mor}(C)} F(\operatorname{cod} f)$ is unique, it follows that $\alpha \gamma=\beta \gamma$. Since $L$ is the object of the equalizer of eq $(\gamma, \beta)$, there must exist a unique factorization $u: M \rightarrow L$ such that $\ell u=\gamma$. Therefore, one has

$$
p_{D} u=\left(r_{D} \ell\right) u=r_{D}(\ell u)=r_{D} \gamma=h_{D},
$$

showing that $u$ is a factorization of of the cone $M$ via the cone $L$ on $F$.
We must show that the factorization $u$ is unique, so that $L$ is the limit of $F$. Suppose $v: M \rightarrow L$ is another factorization, that is, $p_{D} v=h_{D}$ for all $D \in D$. Notice that

$$
r_{D}(\ell u)=\left(r_{D} \ell\right) u=p_{D} u=h_{D}=p_{D} v=\left(r_{D} \ell\right) v=r_{D}(\ell v)
$$

holds for all $D \in D$, therefore both $\ell u$ and $\ell v$ are parallel factorizations-which by Proposition 2.5.4 implies $\ell u=\ell v$. From Proposition 2.3 .5 we know that $\ell$ is monic, therefore $u=v$. This shows the uniqueness of $u$, thus we may conclude that

$$
\lim F=\left(L,\left(p_{D}\right)_{D \in \mathrm{D}}\right)
$$

Proposition 2.5.12 (Finite completeness). Let C be a category. The following properties are equivalent:
(a) The category C is finitely complete.
(b) The category C has a terminal object, binary products and equalizers.
(c) The category C has a terminal object, and pullbacks.

Proof. - If C is finitely complete, then from definition items (b) and (c) hold.

- If (b) holds, then by Proposition 2.1.3 we find that any finite collection has a product. Since $C$ has equalizers, by Theorem 2.5.11 and taking the particular case where $D$ is a finite category, we conclude that $C$ is finitely complete, therefore (b) implies (a).
- If (c) holds, let $1 \in \mathrm{C}$ be the terminal object. For any two objects $A, B \in \mathrm{C}$, the pullback

is the product of $A$ with $B$ in $C$-thus binary products exist. Now consider two parallel morphisms $f, g: A \rightrightarrows B$ in $C$. In this case, one can consider the pullback


Then for $E$ to be the equalizer of $(f, g)$ it suffices to show that $k=\ell$ and $f k=g k$. For the former, notice that

$$
k=\pi_{A} \circ\left(\operatorname{id}_{A} \times f\right) \circ k=\pi_{A} \circ\left(\operatorname{id}_{A} \times g\right) \circ \ell=\ell .
$$

For the last equality, we proceed with a similar argument

$$
f k=\pi_{B} \circ\left(\operatorname{id}_{A} \times f\right) \circ k=\pi_{B} \circ\left(\operatorname{id}_{A} \times g\right) \circ \ell g \ell=g k .
$$

Therefore $\mathrm{eq}(f, g)=(E, k)$, and C has equalizers. This proves that (c) implies (b).

Proposition 2.5.13. Let $F: \mathrm{D} \rightarrow \mathrm{C}$ be a functor, and suppose there exists a collection $\left(f_{j}\right)_{j \in J}$ of morphisms of D generating every other morphism of D -that is, if $g$ is a morphism in D , then $g$ can be written as the composition of finitely many arrows of $\left(f_{j}\right)_{j \in J}$. A cone on the functor $F$ is a pair $\left(L,\left(p_{D}: L \rightarrow F D\right)_{D \in D}\right)$ such that, for any morphism $f_{j}: D \rightarrow D^{\prime}$, we have $p_{D^{\prime}}=F f_{j} \circ p_{D}$.
Proof. To see that this is equivalent to the definition of a cone is simple since, given any $g: D \rightarrow \mathrm{D}^{\prime}$ in D , one has $g=f_{j_{n}} \ldots f_{j_{1}}$ for a finite collection of maps $f_{j_{i}} \in\left(f_{j}\right)_{j \in J}$ where $f_{j i}: D_{i-1} \rightarrow D_{i}$ with $D_{0}=D$ and $D_{n}=D^{\prime}$. Then

$$
\begin{aligned}
F g \circ p_{D} & =F\left(f_{j_{n}} \ldots f_{j_{1}}\right) \circ p_{D} \\
& =\left(F f_{j_{n}} \circ \cdots \circ F f_{j_{1}}\right) \circ p_{D} \\
& =\left(F f_{j_{n}} \circ \cdots \circ F f f_{j_{2}}\right) \circ\left(F f_{j_{1}} \circ p_{D}\right) \\
& =\left(F f_{j_{n}} \circ \cdots \circ F f_{j_{2}}\right) \circ p_{D_{1}} \\
& =\ldots \\
& =F f_{j_{n}} \circ p_{D_{n-1}} \\
& =p_{D^{\prime}},
\end{aligned}
$$

which proves the equivalence.
Definition 2.5.14 (Finitely generated categories). A category C is said to be finitely generated if C is comprised of finitely many objects, and there exists a finite set of morphisms $\left\{f_{1}, \ldots, f_{n}\right\}$ of C such that every morphism of C can be written as a composition of finitely many arrows $f_{j}$.
Proposition 2.5.15. Let D be finitely generated and C be finitely complete, then any functor $F: D \rightarrow C$ has a limit.

Proof. Since C is finitely complete, we can recycle the proof of Theorem 2.5 .11 replacing the pair of products with the finite products $\prod_{D \in D} F D$ and $\prod_{j=1}^{n} F\left(\operatorname{cod} f_{j}\right)$.

## Limit Preserving Functors

Definition 2.5.16 (Limit preserving functor). A functor $F: B \rightarrow C$ is said to preserve limits if-for every small category A and functor $\mathrm{G}: \mathrm{A} \rightarrow \mathrm{B}$-the limit of $G$ exists and the limit of $F G: \mathrm{A} \rightarrow \mathrm{C}$ exists and is given by

$$
\lim (F G)=F(\lim G)
$$

To put more concretely, if $\left(L,\left(p_{A}\right)_{A \in A}\right)$ is the limit of $G$, then the limit of $F G$ is given by $\left(F L,\left(F p_{A}\right)_{A \in \mathrm{~A}}\right)$.

Proposition 2.5.17. Let D be a (finitely) complete category and C be any category. A functor $F: D \rightarrow C$ preserves (finite) limits if and only if it preserves (finite) products and equalizers.

Proof. Suppose F preserves (finite) limits, then in particular it preserves both (finite) products and equalizers. Now, we assume that $F$ preserves both (finite) products and equalizers. Since D is (finitely) complete, for any small (finite) category A and functor $G: A \rightarrow D$, there exists $\lim G$ in $D$. Moreover, since $\lim G$ can be build out of products and equalizers (see Theorem 2.5.11), it follows that, since $F$ preserves products and equalizers, it must also preserve the limit of $G$-and $\lim (F G)=F(\lim G)$.

Proposition 2.5.18. A functor that preserves pullbacks does preserve monomorphisms.
Proof. Let $F: \mathrm{B} \rightarrow \mathrm{C}$ be a functor preserving pullbacks, and consider any monomorphism $f: B \rightarrow B^{\prime}$ in $B$. From Proposition 2.4.7 we find that ker $f=\left(B, \mathrm{id}_{B}, \mathrm{id}_{B}\right)$. Define A to be a category whose objects are $B$ and $B^{\prime}$, and morphisms are identities together with $f: B \rightarrow B^{\prime}$. The limit of the inclusion functor $I: \mathrm{A} \rightarrow \mathrm{D}$ is simply the pullback of $f$ with itself, namely, $\operatorname{ker} f$. Since $F$ preserves pullbacks, then $F I: \mathrm{A} \rightarrow C$ has a limit $\lim (F I)=F$ ker $f$, which is the the pullback of $F I f=F f: F B \rightarrow F B^{\prime}$. This implies in $\operatorname{ker}(F f)=\left(F B, \mathrm{id}_{F B}, \mathrm{id}_{F B}\right)$, which is equivalent to $F f$ being a monomorphism in C . Therefore $F$ preserves monomorphisms.

Proposition 2.5.19. Let $C$ be a category and $C \in C$ be any object. The covariant functor $\operatorname{Mor}_{C}(C,-): \mathrm{C} \rightarrow$ Set preserves all existing limits, including large ones. In particular it preserves monomorphisms.

Proof. Let $F: \mathrm{D} \rightarrow \mathrm{C}$ be a functor admitting a limit $\left(L,\left(p_{D}\right)_{D \in \mathrm{D}}\right)$. Our goal is to show that the functor $\operatorname{Mor}_{C}(C, F(-)): D \rightarrow$ Set—mapping $D \mapsto \operatorname{Mor}_{C}(C, F D)$ and morphisms $\left(f: D \rightarrow D^{\prime}\right) \mapsto\left((F f)_{*}: \operatorname{Mor}_{C}(C, F D) \rightarrow \operatorname{Mor}_{C}\left(C, F D^{\prime}\right)\right)$-has a limit

$$
\lim \operatorname{Mor}(C, F(-))=\left(\operatorname{Mor}_{C}(C, L),\left(\operatorname{Mor}_{C}\left(C, p_{D}\right)\right)_{D \in \mathrm{D}}\right) .
$$

Consider any cone $\left(q_{D}: S \rightarrow \operatorname{Mor}_{C}(C, F D)\right)_{D \in D}$, on Set, over the said composite functor. Notice that every element $s \in S$ induces a cone over $F$ given by $\left(q_{D}(s): C \rightarrow F D\right)_{D \in \mathrm{D}}$. Since $F$ has a limit, there must exist a unique morphism $q_{s}: C \rightarrow L$ in $C$ such that $p_{D} q_{s}=q_{D}(s)$ for all objects $D \in D$. Since $s \in S$ was chosen arbitrarily, we can define a uniquely induced set-function $q: S \rightarrow \operatorname{Mor}_{C}(C, L)$ given by $q(s):=q_{s}$. By the properties of $q_{s}$, the map $q$ also satisfies $\left(F p_{D}\right)_{*} \circ q=q_{D}$ for all $D \in \mathrm{D}$.

Corollary 2.5.20. Given a category $C$ and an object $C \in C$, the contravariant functor Mor $_{C}(-, C): C \rightarrow$ Set transforms existing colimits into limits. In particular, it transforms epimorphisms into monomorphisms.

## Limit Reflection

Definition 2.5.21. A functor $F: B \rightarrow C$ is said to reflect limits if-for all small categories A, and all functors $G: \mathrm{A} \rightarrow \mathrm{B}$, and every cone $\left(L,\left(g_{A}\right)_{A \in \mathrm{~A}}\right)$ over $G —$ the limit of $F G: \mathrm{A} \rightarrow \mathrm{C}$ is given by $\lim (F G)=\left(F L,\left(F g_{A}\right)_{\mathrm{A}}\right)$, then $\lim G=\left(L,\left(g_{A}\right)_{A \in \mathrm{~A}}\right)$.

Proposition 2.5.22. Let $F: \mathrm{A} \rightarrow \mathrm{B}$ be a limit preserving functor. If A is complete, and $F$ reflects isomorphisms, then $F$ reflects limits.

Proof. Let $G: \mathrm{D} \rightarrow \mathrm{A}$ be a functor with $\lim G=\left(L,\left(g_{D}\right)_{D \in \mathrm{D}}\right)$. Since $F$ preserves limits, then

$$
\lim (F G)=\left(F L,\left(F g_{D}\right)_{D \in \mathrm{D}}\right)
$$

Suppose that $\left(F M,\left(F f_{D}\right)_{D \in \mathrm{D}}\right)$ is also a limit of the composite $F G$. We want to show that it must be the case that $\left(L,\left(g_{D}\right)_{D \in \mathrm{D}}\right) \simeq\left(M,\left(f_{D}\right)_{D \in \mathrm{D}}\right)$.

From the limit property in A , there exists a unique factorization $u: M \rightarrow L$ such that for all $D \in \mathrm{D}$ we have $g_{D} u=f_{D}$. Notice that $F u: F M \rightarrow F L$ is a factorization between limits of the composite $F G$. From the uniqueness of the factorization and since $F M \simeq$ $F L$ in B, it follows that $F u$ is an isomorphism. Since $F$ reflects isomorphisms, it must be the case that $u$ is an isomorphism and hence $M \simeq L$, showing that $\left(L,\left(g_{D}\right)_{D \in \mathrm{D}}\right) \simeq$ $\left(M,\left(f_{D}\right)_{D \in \mathrm{D}}\right)$ via $u$.

Proposition 2.5.23. Let $F: \mathrm{A} \rightarrow \mathrm{B}$ be a functor between finitely generated complete categories A and B. If the functor F preserves (or reflects) finite limits, then F preserves (or reflects) finitely generated limits.

Proof. By Proposition 2.5 .15 we have that finitely generated limits can be expressed as finite products and equalizers. Therefore $F$ preserves (or reflects) all existent finitely generated limits.

Proposition 2.5.24. A fully faithful functor $F: C \rightarrow D$ reflects limits.
Proof. Let $G: \mathrm{E} \rightarrow \mathrm{C}$ be any functor from a small category E . Since $F$ is fully faithful, there exists a unique collection of morphisms $\left(g_{E}: L \rightarrow G E\right)_{E \in E}$-which is a cone over $G-s u c h$ that the collection $\left(F g_{E}: F L \rightarrow F G E\right)_{E \in E}$ is a limit cone over $F G$. From the limit property, if $\left(f_{E}: M \rightarrow G E\right)_{E \in E}$ is another cone over $G$, the corresponding cone $\left(F f_{E}: F M \rightarrow F G E\right)_{E \in E}$ over $F G$ has a unique factorization $u: F M \rightarrow F L$ in $D$ such that for all $E \in \mathrm{E}$ we have $F g_{E} \circ u=F f_{E}$. From the fully faithfulness of $F$, there exists a unique morphism $\ell: M \rightarrow L$ in E such that $F \ell=u$-which in this case satisfies $g_{E} \ell=f_{E}$, showing that $\ell$ is the unique factorization between the cones $\left(g_{E}\right)_{E \in E}$ and $\left(f_{E}\right)_{E \in E}$. This implies in $\lim G=\left(L,\left(g_{E}\right)_{E \in E}\right)$ as wanted.

## Absolute (Co)Limits

Definition 2.5.25 (Absolute (co)limit). Let $G: \mathrm{A} \rightarrow \mathrm{B}$ be a functor admitting a colimit colim $G$. We say that such colimit is absolute if, for every functor $F: B \rightarrow C$, we have

$$
\operatorname{colim}(F G)=F(\operatorname{colim} G)
$$

Proposition 2.5.26. Let C be a category and consider the following commutative diagram in C:

$$
A \xrightarrow[g]{\stackrel{f}{\Longrightarrow}} B \xrightarrow{q} C
$$

that is, $q f=q g$. If $g$ and $q$ are both split epimorphisms with sections $s_{g}$ and $s_{q}$, respectively, and the diagram

commutes in C-that is, $s_{q} q=f s_{g}$-then

$$
\operatorname{coeq}(f, g)=(C, q)
$$

Moreover, this coequalizer is absolute.
Proof. Let $p: B \rightarrow D$ be any morphism in $C$ such that $p f=p g$. Define a morphism $\ell:=p s_{q}: C \rightarrow D$, then

$$
\ell q=\left(p s_{q}\right) q=p\left(s_{q} q\right)=p\left(f s_{g}\right)=(p f) s_{g}=(p g) s_{g}=p\left(g s_{g}\right)=p \mathrm{id}_{A}=p
$$

To show the uniqueness of $\ell$, suppose $k: C \rightarrow D$ satisfies $k q=p$. Then we have

$$
k=k\left(q s_{q}\right)=(k q) s_{q}=p s_{q}=\ell,
$$

which proves that coeq $(f, g)=(C, q)$. Since any functor preserves retracts and sections, it follows that the coequalizer is preserved under any given functor.

Proposition 2.5.27 (Absolute pushout). Let C be a category, and consider the following commutative square in C :


If $g$ and $r$ are both split epimorphisms with corresponding sections $\alpha$ and $\beta$ that are compatible with the square-meaning, $f \alpha=\beta h$ - then the square is an absolute pushout.

Proof. We first show that the square is indeed a pushout. Let $X \in \mathrm{C}$ be any object and consider morphisms $p: B \rightarrow X$ and $q: C \rightarrow X$ such that $p f=q g$. Define a morphism $\ell: D \rightarrow X$ by $\ell:=p \beta$. Note that

$$
\begin{gathered}
\ell h=(p \beta) h=p(\beta h)=p(f \alpha)=(p f) \alpha=(q g) \alpha=q(g \alpha)=q \operatorname{id}_{A}=q, \\
\ell r=(p \beta) r=p(\beta r)=p \operatorname{id}_{B}=p .
\end{gathered}
$$

To show uniqueness of $\ell$, suppose $k: D \rightarrow X$ is another morphism such that $k h=q$ and $k r=p$-then in particular $k r=\ell r$ but since $r$ is an epimorphism it follows that $k=\ell$. With this we have proved that the square is indeed a pushout. Since split epimorphisms and commutative diagrams are preserved by any functor, it follows that the pushout is absolute.

## Chapter 3

## Adjoint Functors

### 3.1 Reflections

Definition 3.1.1 (Reflection along a functor). Let $F: C \rightarrow D$ be any functor and $D \in D$. We define a reflection of $D$ along $F$ to be a pair $\left(R_{D}, \eta_{D}\right)$ where $R_{D} \in C$ is an object, and $\eta_{D}: D \rightarrow F R_{D}$ is a morphism of $D$ such that: if $C \in C$ is any object and $\delta: D \rightarrow F C$ is a morphism of D , then there exists a unique morphism $\varepsilon: R_{D} \rightarrow C$ in C such that the diagram

commutes in the category D.
Proposition 3.1.2 (Uniqueness of reflections). Let $F: \mathrm{C} \rightarrow \mathrm{D}$ be a functor and $D \in \mathrm{D}$ be an object. If the reflection of $D$ along $F$ exists, then it is unique up to isomorphism.

Proof. Suppose the reflection of $D$ along $F$ exists and let $\left(R_{D}, \eta_{D}\right)$ and $\left(R_{D}^{\prime}, \eta_{D}^{\prime}\right)$ be two such reflections. From definition:

- Choose the object $R_{D} \in \mathrm{C}$ and the morphism $\eta_{D}: D \rightarrow F R_{D}$ in D. Since $\left(R_{D}^{\prime}, \eta_{D}^{\prime}\right)$ is a reflection, there exists a unique $\varepsilon^{\prime}: R_{D}^{\prime} \rightarrow R_{D}$ such that $F\left(\varepsilon^{\prime}\right) \eta_{D}^{\prime}=\eta_{D}$.
- Analogously, choose the object $R_{D}^{\prime} \in C$ and the morphism $\eta_{D}^{\prime}: D \rightarrow F R_{D}^{\prime}$ in D. Since $\left(R_{D}, \eta_{D}\right)$ is a reflection, there exists a unique $\varepsilon: R_{D} \rightarrow R_{D}^{\prime}$ such that $F(\varepsilon) \eta_{D}=\eta_{D}^{\prime}$.

With $\varepsilon$ and $\varepsilon^{\prime}$ in hands, notice that

$$
F\left(\varepsilon \varepsilon^{\prime}\right) \eta_{D}^{\prime}=F(\varepsilon) F\left(\varepsilon^{\prime}\right) \eta_{D}^{\prime}=F(\varepsilon) \eta_{D}=\eta_{D}^{\prime}=\operatorname{id}_{F R_{D}^{\prime}} \eta_{D}^{\prime}=F\left(\mathrm{id}_{R_{D}^{\prime}}\right) \eta_{D}^{\prime}
$$

and since the composition $\varepsilon \varepsilon^{\prime}$ is unique satisfying such equation, it must be the case that $\varepsilon \varepsilon^{\prime}=\operatorname{id}_{R_{D}^{\prime}}$. Analogously we obtain that $\varepsilon^{\prime} \varepsilon=\operatorname{id}_{R_{D}}$, showing that $R_{D} \simeq R_{D}^{\prime}$.

Proposition 3.1.3. Let $F: C \rightarrow D$ be a functor and assume that for any $D \in D$ the reflection of $D$ along $F$ exists and is $\left(R_{D}, \eta_{D}\right)$. Then there exists a unique functor $R: D \rightarrow C$ such that the following two properties are satisfied:
(a) For any $D \in D$ we have $R D=R_{D}$.
(b) The map $\eta: \mathrm{id}_{\mathrm{D}} \Rightarrow F R$ given by $\left(\eta_{D}: D \rightarrow F R D\right)_{D \in \mathrm{D}}$ is a natural transformation.

Proof. Let $R: \mathrm{D} \rightarrow \mathrm{C}$ be defined as above—we want to prove that $R$ is indeed a functor. Lets define how $R$ acts on Mor(D). Consider a morphism $\delta: D \rightarrow D^{\prime}$ in $D$. The reflection pair $\left(R_{D}, \eta_{D}\right)$ says that there exists a unique morphism $\varepsilon: R_{D} \rightarrow R_{D^{\prime}}$ such that the following diagram commutes


We shall define $R \delta:=\varepsilon$. If $\delta^{\prime}: D^{\prime} \rightarrow D^{\prime \prime}$ is yet another morphism in $D$, then from construction one has

$$
F\left(R\left(\delta^{\prime}\right) R(\delta)\right) \eta_{D}=F R\left(\delta^{\prime}\right) F R(\delta) \eta_{D}=F\left(\varepsilon^{\prime}\right) F(\varepsilon) \eta_{D}=F\left(\varepsilon^{\prime}\right) \eta_{D^{\prime}} \delta=\eta_{D^{\prime \prime}} \delta^{\prime} \delta
$$

On the other hand we have

$$
F R\left(\delta^{\prime} \delta\right) \eta_{D}=\eta_{D^{\prime \prime}}\left(\delta^{\prime} \delta\right)
$$

therefore, from uniqueness we obtain $R\left(\delta^{\prime}\right) R(\delta)=R\left(\delta^{\prime} \delta\right)$.
Definition 3.1.4 (Left adjoint). A functor $R: D \rightarrow C$ is a left adjoint of a functor $F: \mathrm{C} \rightarrow \mathrm{D}$ if there exists a natural transformation $\eta: \mathrm{id}_{\mathrm{D}} \Rightarrow F R$ for which $\left(R D, \eta_{D}\right)$ is a reflection of $D$ along $F$ for any $D \in D$.

Definition 3.1.5 (Coreflection). Let $F: C \rightarrow D$ be a functor and $D \in D$ be any object. A pair $\left(R_{D}, v_{D}\right)$ is said to be a coreflection of $D$ along $F$ if $R_{D} \in C$ and $v_{D}: F R_{D} \rightarrow D$ is a morphism of $D$ such that: for any $C \in C$ and morphism $\delta: F C \rightarrow D$ in $D$ there exists a unique morphism $\varepsilon: C \rightarrow R_{D}$ for which the triangle

commutes in D.
Definition 3.1.6 (Right adjoint). A functor $R: D \rightarrow C$ is a right adjoint of a functor $F: C \rightarrow D$ if there exists a natural transformation $v: F R \Rightarrow \mathrm{id}_{\mathrm{D}}$ such that: for each $D \in \mathrm{D}$ the pair $\left(R D, v_{D}\right)$ is a coreflection of $D$ along $F$.

Theorem 3.1.7 (Equivalent definitions). Let $F: C \rightleftarrows D: G$ be two functors. The following properties are equivalent:
(a) The functor $G$ is left adjoint to $F$.
(b) There exists a pair of natural transformations $\eta$ : $\mathrm{id}_{\mathrm{D}} \Rightarrow F G$ and $v$ : $G F \Rightarrow \mathrm{id}_{\mathrm{C}}$ for which

$$
(F v)(\eta F)=\mathrm{id}_{F}, \text { and }(v G)(G \eta)=\mathrm{id}_{G} .
$$

(c) For every pair of objects $C \in C$ and $D \in D$ there exists a natural bijection

$$
\theta_{C D}: \operatorname{Mor}_{C}(G D, C) \xrightarrow{\simeq} \operatorname{Mor}_{D}(D, F C) .
$$

The naturality of $\theta$ over the choice of $C$ and $D$ can be expressed as follows: for any two other objects $C^{\prime} \in C$ and $D^{\prime} \in D$ together with morphisms $c \in \operatorname{Mor}_{C}\left(C^{\prime}, C\right)$ and $d \in \operatorname{Mor}_{\mathrm{D}}\left(D^{\prime}, D\right)$ the following two squares commute

(d) The functor $F$ is right adjoint to $G$.

Notation 3.1.8. Let $F: C \rightleftarrows D: G$ be functors forming a pair $(G, F)$ of adjoint functors. We shall denote that $G$ is a left adjoint of $F$ (or, which is equivalent, $F$ is right adjoint of $G$ ) by $G \dashv F$.

## Chapter 4

## Abelian Categories

## $4.1 k$-Linear Categories

Definition 4.1.1 ( $k$-linear category). Let $k$ be a commutative ring. A category C is said to be $k$-linear if:
(a) For any pair of objects $X, Y \in \mathrm{C}$ the morphism collection $\operatorname{Mor}_{\mathrm{C}}(X, Y)$ is a $k$-module.
(b) Consider the following diagram in C :

$$
X \xrightarrow{u} A \xrightarrow[g]{\stackrel{f}{\rightrightarrows}} B \xrightarrow{v} Y
$$

Then the following relations are satisfied, which mimic the familiar distributive: given any two scalars $r, t \in k$ we have equalities

$$
\begin{aligned}
& v(r f+t g)=r(v f)+t(v g) \\
& (r f+t g) u=r(f u)+t(g u) .
\end{aligned}
$$

(c) The category C admits finite products and finite coproducts.

A category D is said to be an additive category if D is Z-linear.
Corollary 4.1.2 (Initial and final objects). If C is a $k$-linear category, then C has an initial and final object.

Proof. Since C has all finite products and coproduct, the product $F$ of the empty family of objects of $C$ is the final object of $C$, and the coproduct $I$ of the empty family is the initial object of $C$.

Definition 4.1.3 ( $k$-linear functor). Given $k$-categories $C$ and $D$, we define a $k$-linear functor between $\mathrm{C} \rightarrow \mathrm{D}$ to be a functor $F: \mathrm{C} \rightarrow \mathrm{D}$ such that for all parallel morphisms $f, g: A \rightrightarrows B$ and scalars $r, t \in k$ we have

$$
F(r f+t g)=r(F f)+t(F g)
$$

for any $A, B \in C$. In other words, the induced map $\operatorname{Mor}_{C}(A, B) \rightarrow \operatorname{Mor}_{\mathrm{D}}(F A, F B)$ is a morphism of abelian groups.

Corollary 4.1.4. If C is a $k$-linear category for some commutative ring $k$ then $\mathrm{C}^{\mathrm{op}}$ is $k$-linear.

Definition 4.1.5 (Biproduct). Let $\left(X_{1}, \ldots, X_{n}\right)$ be a finite family of objects of a $k$-linear category C. A pair $\left(X,\left(p_{j}, q_{j}\right)_{j=1}^{n}\right)$ is said to be a biproduct of the family $\left(X_{j}\right)_{j}$ if:
(a) $X$ is an object of $C$.
(b) $p_{j}: X \rightarrow X_{j}$ and $q_{j}: X_{j} \rightarrow X$ are morphisms in C for each $1 \leqslant j \leqslant n$.
(c) The pairs $\left(p_{j}, q_{j}\right)_{j}$ satisfy the following rules:
(1) $\sum_{j=1}^{n} q_{j} p_{j}=\mathrm{id}_{X}$.
(2) For each $1 \leqslant i \neq j \leqslant n$ we have $p_{i} q_{j}=0$, and on the other hand $p_{j} q_{j}=\operatorname{id}_{X_{j}}$.

In other words, $X$ is both a product and a coproduct of the family in a compatible way.
Theorem 4.1.6. Let $\left(X_{1}, \ldots, X_{n}\right)$ be a finite family of objects of a $k$-linear category C .
(a) A pair $\left(P,\left(p_{j}: P \rightarrow X_{j}\right)_{j=1}^{n}\right)$ is a product of $\left(X_{j}\right)_{j}$ if and only if there exists a family of morphisms $\left(q_{j}: X_{j} \rightarrow P\right)_{j=1}^{n}$ in $C$ such that $\left(P,\left(p_{j}, q_{j}\right)_{j}\right)$ is the biproduct of $\left(X_{j}\right)_{j}$.
(b) A pair $\left(C,\left(q_{j}: X_{j} \rightarrow X\right)_{j=1}^{n}\right)$ is a coproduct of the family $\left(X_{j}\right)_{j}$ if and only if there exists a collection of morphisms $\left(p_{j}: C \rightarrow X_{j}\right)_{j=1}^{n}$ for which the pair $\left(C,\left(p_{j}, q_{j}\right)_{j}\right)$ is the biproduct of $\left(X_{j}\right)_{j}$.

Proof. We shall prove only the first item since the second is merely its dual. Suppose that $P$ is a product, and define morphisms $\delta_{j i}: X_{i} \rightarrow X_{j}$ for each pair $1 \leqslant i, j \leqslant n$ where

$$
\delta_{j i}= \begin{cases}\mathrm{id}_{X_{j}}, & \text { if } i=j \\ 0, & \text { if } i \neq j\end{cases}
$$

By the product universal property, for each $1 \leqslant i \leqslant n$ there exists a unique morphism $q_{i}: X_{i} \rightarrow P$ making the following diagram commute for all $1 \leqslant j \leqslant n$ :


Now that we have constructed a family $\left(q_{j}: X_{j} \rightarrow P\right)_{j}$ notice that for each $1 \leqslant j \leqslant n$ one has

$$
p_{j} \sum_{i=1}^{n} q_{i} p_{i}=\sum_{i=1}^{n} p_{j}\left(q_{i} p_{i}\right)=\sum_{i=1}^{n}\left(p_{j} q_{i}\right) p_{i}=\sum_{i=1}^{n} \delta_{j i} p_{i}=p_{j}=p_{j} \operatorname{id}_{P}
$$

Notice that since both $\operatorname{id}_{P}$ and $\sum_{i=1}^{n} q_{i} p_{i}$ satisfy the commutativity property specified above for any $1 \leqslant j \leqslant n$, we can use the universal property of $P$ to obtain that

$$
\sum_{i=1}^{n} q_{i} p_{i}=\operatorname{id}_{P}
$$

Therefore the pair $\left(P,\left(p_{j}, q_{j}\right)_{j}\right)$ satisfies all the requirements to be the biproduct of $\left(X_{j}\right)_{j}$.
Suppose, on the contrary, that there exists a family $\left(q_{j}: X_{j} \rightarrow P\right)_{j}$ of morphisms such that $\left(P,\left(p_{j}, q_{j}\right)_{j}\right)$ is a biproduct of $\left(X_{j}\right)_{j}$. Let $Y \in C$ be any object, and consider a collection of morphisms $\left(f_{j}: Y \rightarrow X_{j}\right)_{j=1}^{n}$. Notice that the map $\sum_{i=1}^{n} q_{i} f_{i}: Y \rightarrow X$ is such that

$$
p_{j} \sum_{i=1}^{n} q_{i} f_{i}=\sum_{i=1}^{n} p_{j}\left(q_{i} f_{i}\right)=\sum_{i=1}^{n}\left(p_{j} q_{i}\right) f_{i}=f_{j}
$$

for each $1 \leqslant j \leqslant n$. Moreover, if $f: Y \rightarrow X$ is another morphism such that $p_{j} f=f_{j}$ for each $j$, then

$$
f=\operatorname{id}_{X} f=\left(\sum_{i=1}^{n} q_{i} p_{i}\right) f=\sum_{i=1}^{n}\left(q_{i} p_{i}\right) f=\sum_{i=1}^{n} q_{i}\left(p_{i} f\right)=\sum_{i=1}^{n} q_{i} f_{i}
$$

This shows that $P$ enjoys the universal property of a product.
Corollary 4.1.7 (Finite products and coproducts are isomorphic). Let C be a $k$-linear category. Every finite collection of objects of C has isomorphic product and coproduct.

Corollary 4.1.8 (Zero object). A $k$-linear category has a zero object.

### 4.2 Properties of $k$-Linear Categories

Definition 4.2.1 (Kernels \& cokernels). Let $f: X \rightarrow Y$ be a morphism in a $k$-linear category C. A kernel of $f$, if it exists, is a pair $(U, u)$ where $U \in C$ and $u: U \rightarrow X$ is a morphism such that
(a) $f u=0$.
(b) For any morphisms $u^{\prime}: U^{\prime} \rightarrow X$ in $C$ such that $f u^{\prime}=0$, there exists a unique morphism $\phi: U^{\prime} \rightarrow U$ such that $u^{\prime}=u g:$


A cokernel of $f$, if it exists, is a pair $(C, c)$ where $C \in C$ and $c: Y \rightarrow C$ such that
(a) $c f=0$.
(b) If $c^{\prime}: Y \rightarrow C^{\prime}$ is a morphism such that $c^{\prime} f=0$, then there exists a unique $\psi: C \rightarrow C^{\prime}$ for which the following diagram commutes:


Corollary 4.2.2 (Kernel is subobject \& cokernel is quotient). Given a morphism $f: X \rightarrow$ $Y$ admitting kernel $(U, u)$ and cokernel $(C, c)$ in a $k$-linear category $C$. Then the isomorphism class of $(U, u)$ is a subobject of $X$, on the other hand, the isomorphism class of $(C, c)$ is a quotient of $Y$.

Proof. Indeed, given a morphism $v: V \rightarrow U$ such that $u v=0$ then $u v=0=u 0$ and by the universal property of the kernel, the unicity of $v$ and 0 yields $v=0$. Let $w: C \rightarrow Z$ be a morphism such that $w c=0$, then $w c=0=0 c$ but then by the universal property of cokernels we obtain $w=0$.

Lemma 4.2.3. In a $k$-linear category C :
(a) A morphism $f$ is monic if and only if $\operatorname{ker} f=0$.
(b) A morphism $g$ is epic if and only if coker $f=0$.

Definition 4.2.4 ((Co)image). Let C be a $k$-category and $f: X \rightarrow Y$ a morphism of C . We define, whenever possible ${ }^{1}$, the following objects:
(a) The image of $f$ is given by $\operatorname{im} f=\operatorname{ker}(\text { coker } f)^{2}$.
(b) The coimage of $f$ is given by coim $f=\operatorname{coker}(\operatorname{ker} f)$.

In other words, we have the following commutative diagram:

which describes the canonical decomposition of $f$.
Definition 4.2.5 (Abelian category). A $k$-linear category is said to be $k$-abelian if:

[^6](a) Every morphism of C admits both kernel and cokernel.
(b) Every morphism $f \in$ C has an associated canonical isomorphism $\bar{f}: \operatorname{coim} f \xlongequal{\simeq} \operatorname{im} f$.

Corollary 4.2.6. A category $C$ is $k$-abelian if and only if the opposite category $C^{\circ}$ is $k$-abelian.

Lemma 4.2.7 (Isomorphisms in abelian categories). In a $k$-abelian category, a morphism is an isomorphism if and only if it is monic and epic.

Proof. It is trivial that an isomorphism is both monic and epic, we shall prove the converse Let $f: X \rightarrow Y$ be both monic and epic in a $k$-abelian category $C$. Since $f$ is monic then $\operatorname{ker} f=0$ and therefore

$$
\operatorname{coim} f=\operatorname{coker}(\operatorname{ker} f)=\operatorname{coker}(0)=\left(X, \operatorname{id}_{X}\right)
$$

On the other hand, since $f$ is epic then coker $f=0$ and thus

$$
\operatorname{im} f=\operatorname{ker}(\operatorname{coker} f)=\operatorname{ker}(0)=\left(Y, \operatorname{id}_{Y}\right)
$$

This shows that $f=\bar{f}$. Now since $\bar{f}$ is an isomorphism due to the fact that C is abelian, then $f$ is an isomorphism.

Theorem 4.2.8. Let C be a $k$-linear category:
(a) If C is such that every morphism admits a kernel. Then every pair of compatible morphisms of C admit a pullback.
(b) If C is such that every morphism admits a cokernel. Then every pair of compatible morphisms of C admit a pushout.

Proof. (a) Let $f: Y \rightarrow X$ and $g: Z \rightarrow X$ be morphisms and let $\pi_{Y}: Y \times Z \rightarrow Y$ and $\pi_{Z}: Y \times Z \rightarrow Z$ be the canonical projections. Let $(U, u)$ be the kernel of the morphism $f \pi_{Y}-g \pi_{Z}: Y \times Z \rightarrow X$. Defining $u_{Y}: U \rightarrow Y$ as $u_{Y}:=\pi_{Y} u$ and $u_{Z}: U \rightarrow Z$ as $u_{Z}:=\pi_{Z} u$, we have from definition that $\left(f \pi_{Y}-g \pi_{Z}\right) u=0$ and therefore $f u_{Y}=g u_{Z}$. Hence the following diagram commutes


To see that $\left(U, u_{Y}, u_{Z}\right)$ is the pullback of $(f, g)$, let $\left(W, w_{Y}: W \rightarrow Y, w_{Z}: W \rightarrow Z\right)$ be a triple in C satisfying $f w_{Y}=g w_{Z}$. From the product universal property we have
a unique morphism $w^{\prime}: W \rightarrow Y \times Z$ such that the following diagram commutes


Therefore

$$
\left(f \pi_{Y}-g \pi_{Z}\right) w^{\prime}=f \pi_{Y} w^{\prime}-g \pi_{Z} w^{\prime}=f w_{Y}-g w_{Z}=0 .
$$

Hence from the universal property of the kernel of $f \pi_{Y}-g \pi_{Z}$ there exists a unique morphism $w: W \rightarrow U$ such that the following diagram commutes:


This shows that the following diagram commutes:

(b) The proof of this second fact becomes dual to item (a) as soon as we use Corollary 4.1.7.

Corollary 4.2.9. In a $k$-abelian category, every pair of compatible morphisms admit a pullback and pushout.

Lemma 4.2.10. Let C be a $k$-abelian category:
(a) Let

be a pullback square in C , then:
(i) If $f_{1}$ is a monomorphism, then so is $p_{2}$.
(ii) If $f_{1}$ is an epimorphism, then so is $p_{2}$.
(b) Let

be a pushout square in $C$, then:
(i) If $g_{2}$ is a monomorphism, then so is $q_{1}$.
(ii) If $g_{2}$ is an epimorphism, then so is $q_{1}$.

Proof. (a) (i) Suppose $f_{1}$ is a monomorphism. Let $u: U \rightarrow P$ be a morphism in C such that $p_{2} u=0$, then

$$
\left(f_{1} p_{1}\right) u=\left(f_{2} p_{2}\right) u=f_{2}\left(p_{2} u\right)=f_{2} 0=0
$$

However, since $f_{1}\left(p_{1} u\right)=0=f_{1} 0$ and $f_{1}$ is monic, then $p_{1} u=0$. This shows that the triple $\left(U, 0: U \rightarrow X_{1}, 0: U \rightarrow X_{2}\right)$ has $u$ as its unique corresponding morphism associated to the universal property of the pullback. Since $u$ is unique and $0: U \rightarrow P$ also satisfies the needed conditions, it follows that $u=0$. This shows that $p_{2} u=0$ implies $u=0$-which is equivalent to $p_{2}$ being monic.
(ii) Suppose $f_{2}$ is an epimorphism. From definition the following sequence is short exact:

$$
0 \longrightarrow P>\quad P \longrightarrow X_{1} \times X_{2} \xrightarrow{f_{1} \pi_{1}-f_{2} \pi_{2}}>X \longrightarrow 0
$$

Let $w: X_{2} \rightarrow W$ be any morphism such that $w p_{2}=0$, and using the fact that $p_{2}=\pi_{2} p$ we find that $w \pi_{2} p=0$. This implies in the existence of a morphism $g: X \rightarrow W$ such that $g\left(f_{1} \pi_{1}-f_{2} \pi_{2}\right)=w \pi_{2}$. On the other hand, if $\iota_{1}: X_{1} \hookrightarrow X_{1} \times X_{2}$ is the canonical inclusion, then

$$
g f_{1}=g\left(f_{1} \pi_{1}-f_{2} \pi_{2}\right) \iota_{1}=w \pi_{2} \iota_{1}=w 0=0
$$

Since $f_{1}$ is assumed to be epic then $g=0$, which in turn implies in $w \pi_{2}=0$. Now using the fact that the canonical projection $\pi_{2}$ is also epic we conclude that $w=0$. This shows that for any morphism with $w p_{2}=0$ one has $w=0$, therefore $p_{2}$ is an epimorphism.
(b) The proof of (b-i) is dual to the proof of (a-ii), while (b-ii) is dual to the proof of (a-i).

Definition 4.2.11 (Exact sequence). Let C be an abelian category and consider a sequence inside C :

$$
\cdots \longrightarrow X_{j+1} \xrightarrow{f_{j+1}} X_{j} \xrightarrow{f_{j}} X_{j-1} \longrightarrow \cdots
$$

Such sequence is said to be a complex if $f_{j} f_{j+1}=0$ for every index $j$. On the other hand, if im $f_{j+1}=\operatorname{ker} f_{j}$ then we say that the sequence is exact in $X_{j}$.

Definition 4.2.12 (Split short exact sequence). Let $C$ be an abelian category. A short exact sequence

$$
0 \longrightarrow X \xrightarrow{f} Y \xrightarrow{g} Z \longrightarrow 0
$$

is split if there exists an isomorphism $Y \simeq X \oplus Z$ such that the following diagram commutes:


Theorem 4.2.13. Let $C$ be an abelian category and consider the following short exact sequence:

$$
0 \longrightarrow X \xrightarrow{f} Y \xrightarrow{g} Z \longrightarrow 0
$$

The following conditions are equivalent:
(a) The sequence is split.
(b) The morphism $f$ is a split monomorphism.
(c) The morphism $g$ is a split epimorphism.

Lemma 4.2.14. Let C be an abelian category and consider the following commutative diagram with short exact rows:


If $\alpha$ and $\gamma$ are isomorphisms, then so is $\beta$.
Theorem 4.2.15. Let C be an abelian category.
(a) Given a diagram in C with an exact bottom row:


Then we can complete this diagram with the pullback $\left(P, p_{1}, p_{2}\right)$ of the pair $\left(f_{1}, f_{2}\right)$ :


Reciprocally, given a commutative diagram with exact rows:


Then there exists an isomorphism $f: P^{\prime} \xrightarrow{\simeq} P$ such that the following diagram commutes:

(b) Given a diagram in C with an exact bottom row:


Then we can complete this diagram with the pushout $\left(Q, q_{1}, q_{2}\right)$ of the pair $\left(g_{1}, g_{2}\right)$ :


Reciprocally, given a commutative diagram with exact rows:


Then there exists an isomorphism $g: P^{\prime} \xrightarrow{\simeq} P$ such that the following diagram commutes:


Proof. We shall only prove item (a) since item (b) follows from duality. Using the universal property for the triple $\left(X_{0}, f_{0}: X_{0} \rightarrow X_{1}, 0: X_{0} \rightarrow X_{2}\right)$ —which satisfies $f_{1} f_{0}=$ $0=f_{2} 0$-we have a unique morphism $p_{0}: X_{0} \rightarrow P$ for which satisfies $p_{1} p_{0}=f_{0}$ and $p_{2} p_{0}=0$. It suffices for us to prove that $p_{0}=\operatorname{ker} p_{2}$ : let $u: U \rightarrow P$ be such that $p_{2} u=0$, then

$$
f_{1} p_{1} u=f_{2} p_{2} u=f_{2} 0=0
$$

Therefore there exists $u^{\prime}: U \rightarrow X_{0}$ such that $f_{0} u^{\prime}=p_{1} u$, moreover, we have

$$
p_{2} p_{0} u^{\prime}=0=p_{2} u \text { and } p_{1} p_{0} u^{\prime}=f_{0} u^{\prime}=p_{1} u .
$$

By the pullback universal property we can use unicity to obtain $u=p_{0} u^{\prime}$. Notice that since $f_{0}=p_{1} p_{0}$ is a monomorphism, then $p_{0}$ is also a monomorphism.

For the reciprocal, notice that the triple ( $P^{\prime}, p_{1}^{\prime}, p_{2}^{\prime}$ ) induces a unique morphism $f: P^{\prime} \rightarrow P$ such that $p_{1} f=p_{1}^{\prime}$ and $p_{2} f=p_{2}^{\prime}$. On the other hand, if we consider the triple ( $X_{0}, f_{0}: X_{0} \rightarrow X_{1}, 0: X_{0} \rightarrow X_{2}$ ) one has that the map $f p_{0}^{\prime}: X_{0} \rightarrow P$ satisfies the commutativity conditions:

$$
\begin{aligned}
& p_{1}\left(f p_{0}^{\prime}\right)=\left(p_{1} f\right) p_{0}^{\prime}=p_{1}^{\prime} p_{0}^{\prime}=f_{0}=p_{1} p_{0} \\
& p_{2}\left(f p_{0}^{\prime}\right)=\left(p_{2} f\right) p_{0}^{\prime}=p_{2}^{\prime} p_{0}^{\prime}=0=p_{2} p_{0}
\end{aligned}
$$

as does the map $p_{0}: X_{0} \rightarrow P$, hence by uniqueness we find that $f p_{0}^{\prime}=p_{0}$. Now by Lemma 4.2.14 we obtain that $f$ is an isomorphism.

### 4.3 Ideals

Definition 4.3.1 (Bilateral ideal). Let C be a $k$-linear category. We define a bilateral ideal of C to be a category $\mathcal{I}$ with $\operatorname{Obj} I \subseteq \operatorname{Obj} C$ and where $I(X, Y) \subseteq \operatorname{Mor}_{C}(X, Y)$ is a submodule for any two $X, Y \in \mathrm{C}$ such that:
(a) If $f \in \mathcal{I}(X, Y)$ and $g \in \operatorname{Mor}_{C}(Y, Z)$ then $g f \in \mathcal{I}(X, Z)$.
(b) If $f \in \mathcal{I}(X, Y)$ and $h \in \operatorname{Mor}_{C}(W, X)$ then $f h \in \mathcal{I}(W, Y)$.

Definition 4.3.2 (Quotient category). Let C be a $k$-linear category and $\mathcal{I}$ be a bilateral ideal of C . We define the quotient category $\mathrm{C} / \mathcal{I}$ to be the $k$-linear category composed of the objects of $C$ and for each $X, Y \in C / I$ we have

$$
\operatorname{Mor}_{C / I}(X, Y):=\operatorname{Mor}_{C}(X, Y) / \mathcal{I}(X, Y)
$$

Definition 4.3.3 (Functor kernel). Let $F: C \rightarrow D$ be a $k$-linear functor. We define the kernel of $F$ to be the bilateral ideal $\operatorname{ker} F$ of $C$ given by the morphisms $f \in$ Mor $C$ such that $F f=0$.

Proposition 4.3.4. Let $F: C \rightarrow D$ be a $k$-linear functor that is both full and essentially surjective. Then there exists a unique equivalence of categories $\bar{F}: C / \operatorname{ker} F \stackrel{\cong}{\leftrightarrows} \mathrm{D}$ such that the following diagram quasi-commutes:

where $P: \mathrm{C} \rightarrow \mathrm{C} / \operatorname{ker} F$ is the canonical projection functor.
Proof. Let $\bar{F}: \mathrm{C} / \operatorname{ker} F \rightarrow \mathrm{D}$ be the functor given by $\bar{F} X:=F X$ for any $X \in \mathrm{C} / \operatorname{ker} F$ and for each morphism $f: X \rightarrow Y$ in $C$ we define

$$
\bar{F}(f+(\operatorname{ker} F)(X, Y))=F f
$$

So that indeed $\bar{F} P=F$. Moreover, it is clear that $\bar{F}$ inherits the properties of being full and essentially surjective from $F$. On the other hand, if $\bar{F}(g+(\operatorname{ker} F)(Z, W))=0$ then $F g=0$ and hence $g \in(\operatorname{ker} F)(Z, W)$, thus $\bar{F}$ is faithful.

### 4.4 Exact Functors

Proposition 4.4.1. Let $F: \mathrm{C} \rightarrow \mathrm{D}$ be a $k$-linear functor between $k$-abelian categories:
(a) If $F$ is covariant, then $F$ is left exact if it preserves kernels—that is, $F(\operatorname{ker} g)=\operatorname{ker}(F g)$ for any $g \in$ Mor $C$. On the other hand, $F$ is said to be right exact if it preserves cokernels-for any $g \in \operatorname{Mor} C$ we have $F(\operatorname{coker} g)=\operatorname{coker}(F g)$.
(b) If $F$ is contravariant, then $F$ is left exact if it transforms kernels into cokernels-that is, $F(\operatorname{ker} g)=\operatorname{coker}(F g)$ for any $g \in \operatorname{Mor} C$. Moreover, $F$ is right exact if it transforms cokernels into kernels-that is, $F(\operatorname{coker} g)=\operatorname{ker}(F g)$ for any $g \in \operatorname{Mor} C$.

## Part II

## Algebra

## Chapter 5

## Vector Spaces

### 5.1 Vector Spaces and Subspaces

## Vector Spaces

Definition 5.1.1 (Vector Space). A set $V$ is called a vector space over a field $k$ (or $k$-space) if it is equipped with an internal operation $+: V \times V \rightarrow V$ where $(a, b) \mapsto a+b$ and the external operation $: k \times V \rightarrow V$ where $(r, a) \mapsto r \cdot a$. We normally call the elements of $V$ as vectors, and the elements of $k$ as scalars. Also, these operations satisfy
I. $(V,+)$ is an abelian group.
II. Multiplication of vectors by scalars is associative and distributive and is unitary (that is, $1 \cdot a=a$ for every $a \in V$ ).

## Subspaces

Definition 5.1.2 (Subspaces). Let $V$ be a vector space. A set $S \subseteq V$ is called a subspace of $V$ if it satisfies all properties of a vector space. Also $S$ is called a proper subspace if it is not equal to the original vector space.

Theorem 5.1.3 (Cover Avoidance). A non-zero vector space $V$ over an infinite field $k$ is not the union of a finite number of proper subspaces.

Proof. Suppose, for the sake of contradiction, that $V=\bigcup_{i \in I} S_{i}$ where $I$ is a finite indexing set and $S_{i}$ are all proper subspaces of $V$. Then, let $S_{1}$ be such that it is not contained in any other subspace $S_{i}$. Define elements $a \in S_{1} \backslash \bigcup_{i \in I \backslash\{1\}} S_{i}$ and $b \in V \backslash S_{1}$ and construct the set $A:=\{r a+b: r \in k\}$. Notice that if we have that one element $r a+b \in S_{1}$, the fact that $a \in S_{1}$ makes $(r a+b)-r a=b \in S_{1}$, contradicting the assumption of $b \notin S_{1}$ thus, we can't have $r a+b \in S_{1}$. Suppose now that we let $r a+b, r^{\prime} a+b \in S_{i}$ different elements, for some $i>1$, then $(r a+b)-\left(r^{\prime} a+b\right)=\left(r-r^{\prime}\right) a \in S_{i}$ but since $k$ is a field, then $a \in S_{i}$ contradicting again the construction and therefore we cannot have more than one element of $A$ contained in $S_{i}$. Now, since $A$ is an infinite set and is contained in $V$, then we cannot have the equality between $V$ and the finite union $\bigcup_{i \in I} S_{i}$.

Definition 5.1.4 (Lattice). A poset $P$ is called a lattice if for every pair of elements of $P$ there exists a join (or a least upper bound) and a meet (greatest lower bound). The set $P$ is called a complete lattice if there exists a join and a meet for every collection of sets and also every collection contains smallest and larger elements under the partial order.

Proposition 5.1.5 (Intersection of subspaces). The intersection of any collection of subspaces of a given vector space is a subspace of the original vector space. This intersection contains the greatest lower bound of subspace that is contained in every subspace of the intersection, then, we denote it by

$$
\bigcap_{i \in I} S_{i}=\operatorname{Glb}\left\{S_{i}: i \in I\right\} .
$$

Proof. Let $V$ be a $k$-space and $S_{i}$ be an arbitrary subspace. Notice that $0 \in \bigcup_{i \in I} S_{i}$. Let elements $u, v \in \bigcap_{i \in I} S_{i}$ so that for all subspace we have that the element $u+v t \in S_{i}$ for any given $t \in k$, thus $u+v t \in \bigcap_{i \in I} S_{i}$, which makes the union a vector space by itself. The claim follows.

Definition 5.1.6 (Sum of subspaces). Let $V$ be a vector space and $S_{i}$ be subspaces of $V$. We define the sum of such subspaces as

$$
\sum_{i \in I} s_{i}:=\left\{\sum_{j} s_{j}: s_{j} \in \bigcup_{i \in I} s_{i}\right\}
$$

Therefore the least upper bound under set inclusion as

$$
\operatorname{Lub}\left\{S_{i}: i \in I\right\}=\sum_{i \in I} S_{i}
$$

Theorem 5.1.7 (Subspaces form a complete lattice). The set containing all subspaces of $V$, denoted by $\mathcal{S}(V)$, is a complete lattice under set inclusion (partial order of the set), with smallest element $\{0\}$ (the zero subspace) and largest element $V$. The meet of any collection of sets $\left\{S_{i}: i \in I\right\}$, where $I$ is a finite indexing set, is

$$
\bigcap_{i \in I} S_{i}=\operatorname{Glb}\left\{S_{i}: i \in I\right\}
$$

and the join is defined as

$$
\sum_{i \in I} S_{i}=\operatorname{Lub}\left\{S_{i}: i \in I\right\}
$$

## Morphisms of Vector Spaces

Definition 5.1.8 (Morphisms). Let $V, L$ be $k$-vector spaces. We say that $\varphi: V \rightarrow L$ is a morphism of vector spaces if it satisfies
I. $\varphi(0)=0$, that is $V \ni 0 \mapsto 0 \in L$.
II. For all $u, v \in V, \varphi(u+b)=\varphi(u)+\varphi(v)$.
III. For all $a \in k$ and $u \in V$ we have $\varphi(a u)=a \varphi(u)$.

Notice in fact that the first property of the morphism is in fact redundant. By means of item two we can see that $\varphi(0)=\varphi(0)+\varphi(0)$, for which the item one is obtained.

Definition 5.1.9. A $k$-linear morphism $f: V \rightarrow V$ is called a linear operator.
Proposition 5.1.10. The $k$-vector spaces, together with morphism between such vector spaces, form a category, of which we'll denote by Vect ${ }_{k}$.
Proof. Certainly, the identity map $\operatorname{id}_{V}: V \rightarrow V$ is a morphism of $k$-vector spaces and all of the properties come directly from the fact that $V$ is a vector space.

Another important feature of morphism between vector spaces is that they are closed under composition. Let $\varphi: V \rightarrow L$ and $\psi: L \rightarrow U$ be morphism between $k$ vector spaces, then the composition $\psi \varphi: V \rightarrow U$ is such that, given any $u, v \in V$ then

$$
\psi(\varphi(u+v))=\psi(\varphi(u)+\varphi(v))=\psi(\varphi(u))+\psi(\varphi(v))
$$

and also, being $a \in k$, we have

$$
\psi(\varphi(a u))=\psi(a \varphi(u))=a \psi(\varphi(u))
$$

Which proves the closure under composition.
We now prove that the composition of morphism of vector spaces is associative. Let the morphisms be as before and define yet another morphism $\ell: W \rightarrow V$, then

$$
(\psi(\varphi \ell))(w)=\psi(\varphi(\ell(w))=(\psi \varphi)(\ell(w))=((\psi \varphi) \ell)(w)
$$

Consider the morphism $\varphi: V \rightarrow L$, then clearly $\varphi \operatorname{id}_{V}=\varphi=\operatorname{id}_{L} \varphi$. Together with the fact that the collection of morphisms between $k$-vector spaces $V, L$ and the collection of morphisms between $k$-vector spaces $U, W$, these two collections are clearly disjoint for $V \neq U$ and $L \neq W$. This finishes the proof of the properties needed for a category.

Proposition 5.1.11 (Initial and Final object). The $k$-vector space 0 is a initial and final object of Vect $_{k}$.

Proof. Essentially, we need to prove that for all $V \in \operatorname{Obj}^{\left(\operatorname{Vect}_{k}\right)}$ there exists unique morphisms $\varphi$ and $\psi$ where


Notice that, since $\varphi$ is a morphism of $k$-vector spaces, we'll need to impose $\varphi(0)=0$ and thus this morphism is clearly unique and satisfies the properties needed for a morphism. Moreover, the morphism has a unique target element, thus the image of $\psi$ is the singleton $\{0\}$ which is also clearly unique and satisfies the properties of morphism.

Definition 5.1.12 (Isomorphism). Let $V, L \in \operatorname{Obj}\left(\operatorname{Vect}_{k}\right)$. We say that $L$ and $V$ are isomorphic, that is $L \simeq V$, if there exists an isomorphism $L \rightarrow V$ in $\operatorname{Mor}(L, V)$.

### 5.2 Matrices

Definition 5.2.1 (Matrix). We define a $m \times n$ matrix with entries in $k$ as morphism $k^{n} \rightarrow k^{m}$ in the category $\operatorname{Vect}_{k}$, that is, regarding $k^{n}, k^{m}$ as $k$-vector spaces.

## Classifying Matrices

Let $e_{j}$ be defined the be a tuple whose $j$-th element is $1 \in k$ and all of the other elements of the tuple are $0 \in k$, moreover, if $e_{j} \in k^{n}$ it is an $n$-tuple. Let the morphism $\varphi \in \operatorname{Mor}_{\mathrm{Vect}_{k}}\left(k^{n}, k^{m}\right)$ and define for all $1 \leqslant j \leqslant n$ the image

$$
\varphi\left(e_{j}\right):=\left(t_{1 j}, t_{2 j}, \ldots, t_{m j}\right)=\sum_{i=1}^{m} t_{i j} e_{i} \in k^{m}
$$

We now prove that in fact the $m n$ elements $t_{i j} \in k$ determine completely the behaviour of $\varphi$, since for any element $\left(a_{j}\right)_{j=1}^{n} \in k^{n}$ we have

$$
\varphi\left(\left(a_{j}\right)_{j=1}^{n}\right)=\varphi\left(\sum_{j=1}^{n} a_{j} e_{j}\right)=\sum_{j=1}^{n} a_{j} \varphi\left(e_{j}\right)=\sum_{j=1}^{n}\left(\sum_{i=1}^{m} a_{j} t_{i j} e_{i}\right) \in k^{m} .
$$

Moreover, it can trivially be seen that the mapping $\left(a_{j}\right)_{j=1}^{n} \mapsto \sum_{j=1}^{n} \sum_{i=1}^{m} a_{j} t_{i j} e_{i}$ is indeed a morphism of vector spaces since, for another $\left(b_{j}\right)_{j=1}^{n} \in k^{n}$, we have

$$
\varphi\left(\left(a_{j}+b_{j}\right)_{j=1}^{n}\right)=\varphi\left(\left(a_{j}\right)_{j=1}^{n}\right)+\varphi\left(\left(b_{j}\right)_{j=1}^{n}\right)
$$

also, given $c \in k$ we have

$$
\varphi\left(\left(c a_{j}\right)_{j=1}^{n}\right)=c \varphi\left(\left(a_{j}\right)_{j=1}^{n}\right)
$$

Since the matrix $k^{n} \rightarrow k^{m}$ can be identified and completely determined with elements $\left(t_{i j}\right)_{i, j}$ where $1 \leqslant i \leqslant m$ and $1 \leqslant j \leqslant n$, then we visually can represent it by

$$
\varphi=\left(\begin{array}{ccc}
t_{11} & \ldots & t_{1 n} \\
\vdots & \ddots & \vdots \\
t_{m 1} & \ldots & t_{m n}
\end{array}\right)
$$

Also, we can regard the fact that $\varphi\left(\left(a_{i}\right)_{i=1}^{n}\right)=\left(b_{i}\right)_{i=1}^{m}$ visually as a system of equations

$$
\left(\begin{array}{ccc}
t_{11} & \ldots & t_{1 n} \\
\vdots & \ddots & \vdots \\
t_{m 1} & \ldots & t_{m n}
\end{array}\right)\left(\begin{array}{c}
a_{1} \\
\vdots \\
a_{n}
\end{array}\right)=\left(\begin{array}{c}
b_{1} \\
\vdots \\
b_{m}
\end{array}\right)
$$

Definition 5.2.2 (Square diagonal Matrix). We define a $n \times n$ matrix $A$ to be diagonal if for all $i \neq j$ indices we have $a_{i j}=0$.

Definition 5.2.3 (Matrix for a linear map). Suppose $V \simeq k^{n}$ and $W \simeq k^{m}$ are $k$-vector spaces and $L: V \rightarrow W$ is a linear morphism. Let also $\left\{v_{j}\right\}_{j=1}^{n}$ and $\left\{w_{i}\right\}_{i=1}^{m}$ be basis for the respective given finite dimensional vector spaces. By the isomorphism, we can represent $L$ as a matrix $k^{n} \rightarrow k^{m}$ whose components $t_{i, j}$ with $1 \leqslant i \leqslant m$ and $1 \leqslant i \leqslant n$ are defined with respect to the given basis as

$$
L v_{j}=\sum_{i=1}^{m} t_{i, j} w_{i}
$$

## Matrix multiplication

Definition 5.2.4 (Multiplication of matrices). Let $A=\left[a_{i, j}\right]: k^{n} \rightarrow k^{m}$ and $B=\left[b_{i, j}\right]: k^{\ell} \rightarrow$ $k^{n}$ be matrices. Then, the product of the matrices $A$ and $B$ is defined as $A B: k^{n} \rightarrow k^{\ell}$ with coefficients

$$
c_{i, j}=\sum_{k=1}^{n} a_{i, k} b_{k, j}
$$

Proposition 5.2.5 (Composition of morphisms matrix). Let $k$-linear morphisms $V \xrightarrow{g}$ $W \xrightarrow{f} L$ of finite $k$-dimensional vector spaces, and choose basis $\left\{v_{j}\right\},\left\{w_{k}\right\},\left\{l_{i}\right\}$ to be basis of $V, W, L$ respectively, and let $A_{g}$ and $A_{f}$ be the matrix representation of the morphisms $g$ and $f$ with respect to the given basis. Then the matrix representation of the composition $f g: V \rightarrow L$ is given by $A_{f g}=A_{f} A_{g}$.

Proof. Let $A_{f}:=\left[a_{i, k}\right]$, and $A_{g}:=\left[b_{k, j}\right]$, and $A_{f g}=\left[c_{i, j}\right]$. Then, since from definition we have $g\left(v_{j}\right)=\sum_{k} b_{k, j} w_{k}$, hence

$$
f g\left(v_{j}\right)=\sum_{k} b_{k, j} f\left(w_{k}\right)=\sum_{k}\left(b_{k, j}\left(\sum_{i} a_{i, k} l_{i}\right)\right)=\sum_{i}\left(\sum_{k} a_{i, k} b_{k, j}\right) l_{i}=\sum_{i} c_{i, j} l_{i}
$$

thus $c_{i, j}=\sum_{k} a_{k} i, k b_{k, j}$ and thus $A_{f g}=A_{f} A_{g}$ as wanted.
Definition 5.2.6 (Conjugation). Let $M_{n}(k)$ be the collection of matrices $k^{n} \rightarrow k^{n}$. We define a matrix conjugation as the linear morphism $M_{n}(k) \rightarrow M_{n}(k)$ with the mapping $A \mapsto B^{-1} A B$, where $B$ is an invertible matrix.

Proposition 5.2.7. Every conjugation is an automorphism of the matrix algebra $M_{n}(k)$.

### 5.3 Product and Direct Sum of Vector Spaces

## Free Vector Spaces

Proposition 5.3.1. Let $S$ be a set and define the power set $k^{S}$ with addition and scalar multiplication, that is, given $f, g \in k^{S}$ and $a \in k$ we have $(f+g)(x)=f(x)+g(x)$ and $(a f)(x)=a f(x)$. Then the set $k^{S}$ is a $k$-vector space.

Proof. Notice that clearly $\left(k^{S},+\right)$ is an abelian group from the construction of the additive structure, moreover, since $k$ is a field, it inherit the associativity and distributivity of scalar multiplication. The 0 vector can be regarded as the map whose image is the singleton $\{0\}$.

Proposition 5.3.2 (Functoriality of $k^{S}$ ). Let sets $S, S^{\prime}$ and a map $\alpha: S \rightarrow S^{\prime}$, then

$$
\alpha^{*}: k^{S^{\prime}} \rightarrow k^{S} \text { mapping }\left(f: S^{\prime} \rightarrow k\right) \mapsto(f \alpha: S \rightarrow k)
$$

is a morphism of $k$-vector spaces. Moreover if $\beta: S^{\prime} \rightarrow S^{\prime \prime}$ then $(\beta \alpha)^{*}=\alpha^{*} \beta^{*}$.
Proof. Let maps $f, g: S^{\prime} \rightarrow k$, then

$$
\left(f+g: S^{\prime} \rightarrow k\right) \stackrel{\alpha^{*}}{\mapsto}((f+g) \alpha: S \rightarrow k)
$$

but for every $s \in S$ we have $(f+g)(\alpha(s))=f(\alpha(s))+g(\alpha(s))$, which shows the first property. For the second property, let $a \in k$, then we get the map

$$
\left((a \cdot f): S^{\prime} \rightarrow k\right) \stackrel{\alpha^{*}}{\longmapsto}((a \cdot f) \alpha: S \rightarrow k),
$$

but for each $s \in S$ we have $(a \cdot f)(\alpha(s))=a \cdot f(\alpha(s))$, proving the second property, which finishes the proof.

Definition 5.3.3 (Free vector space). Let $S$ be a set, we define the free vector space on $S$ to be the object

$$
k^{\oplus S}=\left\{f \in k^{S}: f(s) \neq 0 \text { only for finitely many } s \in S\right\}
$$

together with an additive structure and scalar multiplication, satisfying for all $f, g \in$ $k^{\oplus S}, s \in S$ and $a \in k$ :

$$
(f+g)(s)=f(s)+g(s) \text { and }(a \cdot f)(s)=a \cdot f(s)
$$

Proposition 5.3.4. The object $k^{\oplus S}$ is a $k$-vector space.
Proof. Let the maps $f, g \in k^{\oplus S}$, so that the specification of finiteness of non-zero values is satisfied. Denote $A, B$ the finite sets containing the non-zero elements of $S$ under, respectively, $f$ and $g$. Notice now that the map $f+g \in k^{S}$ is such that for all $s \in A \cup B$ then $(f+g)(s)=f(s)+g(s) \neq 0$ thus, denoting $C$ as the set containing all non-zero elements of $S$ under $f+g$, we see that $A \cup B \subseteq C$. The converse is trivial since if $s \in C$ then $f(s)+g(s) \neq 0$ and thus at least one of the images is non-zero, thus $s \in A \cup B$ and then $C \subseteq A \cup B$, which implies that $A \cup B=C$ and, in particular, $C$ is finite. Therefore $f+g \in k^{\oplus S}$. Note now that for any $c \in \mathbf{R} \backslash\{0\}$, then $c \cdot f$ has non-zero values only in $A \subseteq S$, in the case $c=0$ then $c \cdot f$ is zero all over $S$, which makes $c \cdot f \in k^{\oplus S}$. Thus indeed $k^{\oplus S}$ is a vector space.

For each $s \in S$ we can define a map s: $S \rightarrow k$ such that

$$
\mathbf{s}(x):= \begin{cases}1, & \text { for } x=s \\ 0, & \text { for } x \neq s\end{cases}
$$

Also, this notion comes together with the natural monomorphism

$$
\iota: S \hookrightarrow k^{\oplus S} \text { mapping } s \mapsto \mathbf{s}
$$

which allow us to write any element $v \in k^{\oplus S}$ as a linear combination, with non-zero scalars $a_{i} \in k$, of the form

$$
v=\sum_{i=1}^{n} a_{i} \mathbf{s}_{i} \in k^{\oplus S}, \text { mapping } s \mapsto \begin{cases}a_{i}, & \text { if } s=s_{i} \text { for some } i \\ 0, & \text { if } s \neq s_{i} \text { for all } i\end{cases}
$$

Proposition 5.3.5 (Universal property of the free vector spaces). Let $S$ be a set, and $V$ be a $k$-vector space, and a function $f: S \rightarrow V$. Then there exists a unique $k$-linear morphism $\ell: k^{\oplus S} \rightarrow V$ such that the diagram commutes:


Proof. (Uniqueness) Notice that for all $\alpha \in k^{\oplus S}$ we can write it as a finite sum $\alpha=$ $\sum_{s \in S} \alpha(s) \mathbf{s}$ (the finiteness comes from the fact that only finitely may $\mathbf{s}$ will actually appear, since there are only finite non-zero values under $\alpha$ ). Let now $L: k^{\oplus S} \rightarrow V$ be a morphism of $k$-vector spaces such that $f=L \iota$ for a given $f: S \rightarrow V$. Then

$$
L(\alpha)=\sum_{s \in S} L(\alpha(s) \mathbf{s})=\sum_{s \in S} \alpha(s) L(\mathbf{s})=\sum_{s \in S} \alpha(s)(L \iota)(s)=\sum_{s \in S} \alpha(s) f(s)
$$

which implies that the morphism $L$ is indeed unique.
To prove the existence of $L$, consider its definition as above. Notice that for any $\alpha, \beta \in k^{\oplus S}$ we have

$$
L(\alpha+\beta)=\sum_{s \in S}(\alpha+\beta)(s) f(s)=\sum_{s \in S} \alpha(s) f(s)+\sum_{s \in S} \beta(s) f(s)=L(\alpha)+L(\beta) .
$$

Moreover, if $a \in k$, then $k \alpha=k \sum_{s \in S} \alpha(s) \mathbf{s}$, thus indeed

$$
L(k \alpha)=\sum_{s \in S}(k \alpha)(s) f(s)=k \sum_{s \in S} \alpha(s) f(s)=k L(\alpha)
$$

proving the last condition in order to be a morphism of $k$-vector spaces,

This way we find that actually

$$
\operatorname{Mor}_{\text {Set }}(S, V) \simeq \operatorname{Mor}_{\text {Vect }_{k}}\left(k^{\oplus S}, V\right)
$$

We now want in some way to construct a morphism that maps $k^{\oplus S^{\prime}} \rightarrow k^{\oplus S}$, where $S, S^{\prime}$ are sets, just as we've done for the case $k^{S^{\prime}} \rightarrow k^{S}$. To do so, we can first consider the maps $\alpha: S \rightarrow S^{\prime}$, and the inclusion maps $\iota_{S}: S \rightarrow k^{\oplus S}$ and $\iota_{S^{\prime}}: S^{\prime} \hookrightarrow k^{\oplus S^{\prime}}$, and construct the map

$$
\ell=\iota S^{\prime} \alpha: S \rightarrow k^{\oplus S^{\prime}}
$$

Then, by means of the universal property for free vector spaces, we find that there exists a unique morphism of $k$-vector spaces $\alpha_{*}: k^{\oplus S} \rightarrow k^{\oplus S^{\prime}}$ such that $\ell=\alpha_{*} \iota S$, that is, the following diagram commutes


## Product

Definition 5.3.6 (Pair product). Let $L, V$ be $k$-vector spaces. Then the product

$$
L \times V=\{(\ell, v): \ell \in L, v \in V\}
$$

equipped with pairwise sum and product by scalar is a $k$-vector space.
Definition 5.3.7 (General product). Let $\left\{V_{i}\right\}_{i \in I}$ be a set of $k$-vector spaces, with an indexing set $I$. The product of these vector spaces is defined as

$$
\prod_{i \in I} V_{i}=\left\{\left(v_{i}\right)_{i \in I}: v_{i} \in V_{i}\right\}
$$

equipped with addition and multiplication by scalar as

$$
\left(v_{i}\right)_{i \in I}+\left(u_{i}\right)_{i \in I}=\left(v_{i}+u_{i}\right)_{i \in I} \text { and } a\left(v_{i}\right)_{i \in I}=\left(a v_{i}\right)_{i \in I}
$$

is a $k$-vector space.
Proposition 5.3.8. Let $S$ be a set. Then there exists a natural isomorphism

$$
k^{S} \simeq \prod_{s \in S} k
$$

so that $\prod_{s \in S} k$ generalizes the power set.
Proof. For any $f \in k^{S} \operatorname{map}(f: S \rightarrow k) \mapsto(f(s))_{s \in S}$. Notice that the mapping is obviously a monomorphism, since the maps $S \rightarrow k$ are well defined, and is also an epimorphism because we just need to create a function fitting such an image.

Proposition 5.3.9 (Universal property for products). Let the indexed set of $k$-vector spaces $\left\{V_{i}\right\}_{i \in I}$. For any given $k$-vector space $L$, together with a morphism $\varphi_{j}: L \rightarrow V$, there exists a unique morphism $\ell: L \rightarrow \prod_{i \in I} V_{i}$ such that the diagram commutes:


Proof. Notice that this is simply the product universal product restricted for only one branch of the product $V_{j} \times \prod_{i \neq j} V_{i}$. For the uniqueness, notice that since we want $\varphi_{j}=\pi_{j} \ell$, then it ought to be the case that

$$
L \ni x \stackrel{\ell}{\longmapsto}\left(\varphi_{j}(x)\right)_{i \in I} \in \prod_{i \in I} V_{i}
$$

since the composition needs to be satisfied for any index $j \in I$. Therefore the morphism $\ell$ is indeed unique, if it exists. Now we show its existence. Notice that since $\varphi_{j}$ is a $k$ linear map, then indeed for any $x, y \in L$ and $a \in k$ we have $\ell(x+a y)=\left(\varphi_{i}(x+a y)\right)_{i \in I}=$ $\left(\varphi_{i}(x)\right)_{i \in I}+a\left(\varphi_{i}(y)\right)_{i \in I}=\ell(x)+a \ell(y)$, which shows the linearity and the existence. 呐

Definition 5.3.10 (Direct sum). Given an indexed collection of $k$-vector spaces $\left\{V_{i}\right\}_{i \in I}$, we define the direct sum of them as

$$
\bigoplus_{i \in I} V_{i}:=\left\{\left(v_{i}\right)_{i \in I}: v_{i} \in V_{i}, \text { where } v_{i} \neq 0 \text { finitely many times in } I\right\}
$$

together with addition and scalar multiplication being defined component-wise.
Proposition 5.3.11. The direct sum of $k$-vector spaces is again a $k$-vector space.
Proposition 5.3.12. Let a set $S$. Then, the exists natural isomorphism $\bigoplus_{s \in S} k \simeq k^{\oplus S}$.
Proof. Notice that if we make the morphism $\ell: \bigoplus_{s \in S} k \rightarrow k^{\oplus S}$ defined as the mapping

$$
\bigoplus_{s \in S} k \ni\left(a_{s}\right)_{s \in S} \stackrel{\ell}{\longmapsto}\left(s \stackrel{f}{\mapsto} a_{s}\right) \in k^{\oplus S}
$$

then we see that no information is lost since this is surely injective, because the morphisms of $k^{\oplus S}$ are well defined, and given any morphism $f \in k^{\oplus S}$ we know that its non-zero values are finite, thus we can construct its image as a tuple in $\bigoplus_{s \in S} k$, which allow us to say that such mapping is also surjective. Therefore, the morphism $\ell$ is an isomorphism. We now show that it is indeed $k$-linear. Let the tuples $\left(a_{s}\right)_{s \in S}$ and $\left(b_{s}\right)_{s \in S}$, then

$$
\left(a_{s}\right)_{s}+\left(b_{s}\right)_{s}=\left(a_{s}+b_{s}\right)_{s} \stackrel{\ell}{\mapsto}\left(s \stackrel{f}{\mapsto} a_{s}+b_{s}\right)=\left(s \stackrel{g}{\mapsto} a_{s}\right)+\left(s \stackrel{h}{\mapsto} b_{s}\right)
$$

Moreover, if $c \in k$ is any scalar, then $\left(c a_{s}\right)_{s} \longmapsto\left(s \mapsto c a_{s}\right)=c\left(s \mapsto a_{s}\right)$, which shows the last property for $\ell$ being a $k$-linear morphism.

Proposition 5.3.13 (Universal property for the direct sum). Let the collection of $k$-vector spaces $\left\{V_{i}\right\}_{i \in I}$ and for every $j \in I$ define the inclusion $\iota_{j}: V_{j} \rightarrow \bigoplus_{i \in I} V_{i}$, where

$$
v \stackrel{\iota_{j}}{\longmapsto}\left(v_{i}\right)_{i \in I}, \text { where } v_{i}=\left\{\begin{array}{ll}
v, & i=j \\
0, & i \neq j
\end{array} .\right.
$$

Then, for any arbitrary $k$-vector space $L$, together with $k$-linear morphisms $\left(f_{i}: V_{i} \rightarrow\right.$ $L)_{i \in I}$, there exists a unique $k$-linear morphism $\ell: \bigoplus_{i \in I} V_{i} \rightarrow L$ such that the diagram commutes for every $j \in I$ :


Proof. Since we want to have $\ell_{\iota_{j}}=f_{j}$, it must be the case for any $v \in v_{j}$ that, being $\iota_{j}(v):=\left(v_{i}\right)_{i}$, then

$$
\ell\left(\left(v_{i}\right)_{i \in i}\right)=\ell\left(\sum_{i \in i} \iota_{i}\left(v_{i}\right)\right)=\sum_{i \in i} \ell\left(\iota_{i}\left(v_{i}\right)\right)=\sum_{i \in i} f_{i}\left(v_{i}\right)
$$

thus indeed necessarily, if such a map exists, then $\ell$ is unique. We now check that $\ell$ is indeed a $k$-linear morphism.

$$
\begin{gathered}
\ell\left(\left(v_{i}+w_{i}\right)_{i}\right)=\sum_{i} f_{i}\left(v_{i}+w_{i}\right)=\sum_{i} f_{i}\left(v_{i}\right)+\sum_{i} f_{i}\left(w_{i}\right) \\
\ell\left(\left(a v_{i}\right)_{i}\right)=\sum_{i} f_{i}\left(a v_{i}\right)=a \sum_{i} f_{i}\left(v_{i}\right)
\end{gathered}
$$

Which proves that $\ell$ is indeed a morphism of $k$ vector spaces as wanted.
Proposition 5.3.14. Let $\left\{S_{i}\right\}_{i \in S}$ be a collection of sets. There exists a canonical isomorphism

$$
\bigoplus_{i \in I} k^{\oplus S_{i}} \simeq k^{\oplus\left(\amalg_{i \in I} S_{i}\right)}
$$

Proof. Just use both universal properties for coproduct and free vector spaces, this establishes a two way unique morphism, which proves the canonical isomorphism. $\quad$.

Proposition 5.3.15. Let the $k$-linear morphism $T: k^{n} \rightarrow k^{m}$ and $S: k^{n} \rightarrow k^{\ell}$. If we consider their matrix representation, they induce the $k$-linear morphism

$$
M: k^{n} \rightarrow k^{m+\ell} \text { represented by } M=\binom{T}{S}
$$

Proof. Notice that from the universal property of products we have


Where the isomorphism $k^{m} \times k^{\ell} \simeq k^{m+\ell}$ is induced by the mapping $\left(\left(a_{i}\right)_{i=1,}^{m}\left(b_{i}\right)_{i=1}^{\ell}\right)=$ $\left(c_{i}\right)_{i=1}^{m+\ell}$ where $c_{i}:=a_{i}$ for all $1 \leqslant i \leqslant m$ and $c_{i}:=b_{i}$ for all $m+1 \leqslant i \leqslant m+\ell$. Notice then that the morphism $M$ is induced by the morphism $T, S$ so that $T=\pi_{m} M$ and $S=\pi_{\ell} M$, so that we need to have $M(v)=(T(v), S(v))$ for all $v \in k^{m}$, but notice that $M: k^{n} \rightarrow k^{m} \times k^{\ell}$ is in fact isomorphic to a $k$-linear map $k^{n} \rightarrow k^{m+\ell}$, so that we can encode the transformation of $M$, without losing, information as follows. Suppose that

$$
T=\left[\begin{array}{ccc}
a_{11} & \ldots & a_{1 n} \\
\vdots & \ddots & \vdots \\
a_{m 1} & \ldots & a_{m n}
\end{array}\right] \text { and } S=\left[\begin{array}{ccc}
b_{11} & \ldots & b_{1 n} \\
\vdots & \ddots & \vdots \\
b_{\ell 1} & \ldots & b_{\ell n}
\end{array}\right]
$$

then we can say that $M$ can be isomorphically represented by

$$
M(v)=\left(\left(u_{i}\right)_{i=1}^{m}\left(u_{i}\right)_{i=m+1}^{m+\ell}\right) \simeq\left[\begin{array}{ccc}
a_{11} & \ldots & a_{1 n}  \tag{দ}\\
\vdots & \ddots & \vdots \\
a_{m 1} & \ldots & a_{m n} \\
b_{(m+1) 1} & \ldots & b_{(m+1) n} \\
\vdots & \ddots & \vdots \\
b_{(m+\ell) 1} & \ldots & b_{(m+\ell) n}
\end{array}\right]\left[\begin{array}{c}
v_{1} \\
\vdots \\
v_{n}
\end{array}\right]=\left[\begin{array}{c}
u_{1} \\
\vdots \\
u_{m+\ell}
\end{array}\right]
$$

where $\left(u_{i}\right)_{i=1}^{m}=T(v)$ and $\left(u_{i}\right)_{i=m+1}^{m+\ell}=S(v)$.
Proposition 5.3.16. Let the $k$-linear morphism $T: k^{n} \rightarrow k^{\ell}$ and $S: k^{m} \rightarrow k^{\ell}$. They induce the $k$-linear morphism

$$
M: k^{n+m} \simeq k^{n} \oplus k^{m} \rightarrow k^{\ell} \text { represented by } M=\left(\begin{array}{ll}
T & S
\end{array}\right)
$$

Proof. From the universal property of the coproduct we have that


Where $k^{n}, k^{m}$ are both finite, then $k^{n} \oplus k^{m} \simeq k^{n} \times k^{m} \simeq k^{n+m}$, which proves the isomorphism written. Therefore $M: k^{n} \oplus k^{m} \rightarrow k^{\ell}$ is isomorphic to a $k$-linear morphism
$k^{n+m} \rightarrow k^{\ell}$. Moreover, if $T$ and $S$ are represented by

$$
T=\left[\begin{array}{ccc}
a_{11} & \ldots & a_{1 n} \\
\vdots & \ddots & \vdots \\
a_{\ell 1} & \ldots & a_{\ell n}
\end{array}\right] \text { and } S=\left[\begin{array}{ccc}
b_{11} & \ldots & b_{1 m} \\
\vdots & \ddots & \vdots \\
b_{\ell 1} & \ldots & b_{\ell m}
\end{array}\right]
$$

Then from the construction $T=M \iota_{n}$ and $S=M \iota_{m}$ we conclude that for all element $a=\left(a_{i}\right)_{i=1}^{n} \in k^{n}$ and every $b=\left(b_{i}\right)_{i=1}^{m} \in k^{m}$, will be such that

$$
\begin{aligned}
& k^{n} \ni\left(a_{i}\right)_{i=1}^{n} \stackrel{\iota_{n}}{\longmapsto}\left(a_{1}, \ldots, a_{n}, 0, \ldots, 0\right) \stackrel{M}{\longmapsto}\left(a_{i}^{\prime}\right)_{i=1}^{\ell}=T(a) \in k^{\ell} \\
& k^{m} \ni\left(b_{i}\right)_{i=1}^{m} \stackrel{\iota_{m}}{\longmapsto}\left(b_{1}, \ldots, b_{m}, 0, \ldots, 0\right) \stackrel{M}{\longmapsto}\left(b_{i}^{\prime}\right)_{i=1}^{\ell}=S(b) \in k^{\ell}
\end{aligned}
$$

therefore, if we have the element

$$
\left(a_{1}, \ldots, a_{n}, b_{1}, \ldots, b_{m}\right) \simeq\left(\left(a_{1}, \ldots, a_{n}\right),\left(b_{1}, \ldots, b_{m}\right)\right) \stackrel{M}{\longmapsto}\left(a_{i}^{\prime}+b_{i}^{\prime}\right)_{i=1}^{\ell}=T(a)+S(b)
$$

which shows that indeed

$$
M=\left[\begin{array}{cccccc}
a_{11} & \ldots & a_{1 n} & b_{1(n+1)} & \ldots & b_{1(n+m)} \\
\vdots & \ddots & \vdots & \vdots & \ddots & \vdots \\
a_{\ell 1} & \ldots & a_{\ell n} & b_{\ell(n+1)} & \ldots & b_{\ell(n+m)}
\end{array}\right]=\left[\begin{array}{ll}
T & S
\end{array}\right] .
$$

### 5.4 Quotients and Subspaces

Definition 5.4.1 (Subspace). Let $V$ be $k$-vector space. We say that $W \subseteq V$ is a subspace of $V$ if
I. $0 \in W$.
II. If $v, u \in W$ then $v+u \in W$.
III. For all $c \in k$ and for all $v \in W$, we have $c v \in W$.

Definition 5.4.2 (Quotient). Let $V$ be a $k$-vector space and $W \subseteq V$ a subspace. The quotient of $V$ by $W$ is defined as $V / W=V / \sim$ where $v \sim u \Leftrightarrow v-u \in W$.

Theorem 5.4.3 (Universal property for quotients). Let $V$ be a $k$-vector space and $W \subseteq V$ be a subspace. Let also $L$ be any $k$-vector space and a $k$-linear morphism $f: V \rightarrow L$ such that for all $w \in W$ we have $f(w)=0$, that is $W \subseteq \operatorname{ker} f$. There exists a unique $k$-linear morphism $\ell: V / W \rightarrow L$ such that the following diagram commutes


Moreover, $\operatorname{ker} \ell=\operatorname{ker} f / W$ and $\operatorname{im}(\ell)=\operatorname{im}(f)$.

Proof. Since we need $\ell \pi=f$ then $\ell \pi: v \stackrel{\pi}{\mapsto}[v] \stackrel{\ell}{\mapsto} f(v)$. We first show that it is indeed well defined. For that, suppose $v \sim u$, that is $[v]=[u]$, then we know that since they belong to the same class, they must have the same image under the mapping $\ell$. Notice that since $v \sim u \Leftrightarrow v-u \in W \subseteq \operatorname{ker} f$ then $f(v-u)=f(v)-f(u)=0$ and therefore indeed $f(v)=f(u)$ whenever $[v]=[u]$. The uniqueness of $\ell$ comes from the fact that its image depends strictly on the unique image of $f$.

Now we show the last two propositions. For the first one, notice that ker $\ell=\{[v] \in$ $V / W: v \in \operatorname{ker} f\}=\operatorname{ker} f / W$. For the last statement of the theorem we have that since $\ell \pi=f$ then $\operatorname{im}(f) \subseteq \operatorname{im}(\ell)$ and also $\pi$ is surjective, thus $\operatorname{im}(\ell) \subseteq \operatorname{im}(f)$, which proves what is proposed at last.

Theorem 5.4.4 (First Isomorphism). Let $f: V \rightarrow W$ be a $k$-linear morphism, then the injective $k$-linear morphism $\ell: V / \operatorname{ker} f \hookrightarrow W$ is such that

$$
V / \operatorname{ker} f \simeq \operatorname{im}(f)
$$

is an isomorphism of $k$-vector spaces.
Proof. This is just a special case of the universal property.
Theorem 5.4.5 (Second Isomorphism). Let $V$ be a $k$-vector space and $U, W \subseteq V$ be subspaces. Then there is a natural isomorphism

$$
(W+U) / U \simeq W /(W \cap U)
$$

Proof. Notice that this isomorphism needs to be the mapping

$$
(W+U) / U \ni[w+u] \longmapsto[w] \in W /(W \cap U)
$$

Notice that we have the mappings

$$
\begin{gathered}
{[w+u]+\left[w^{\prime}+u^{\prime}\right]=\left[\left(w+w^{\prime}\right)+\left(u+u^{\prime}\right)\right] \longmapsto\left[w+w^{\prime}\right]=[w]+\left[w^{\prime}\right]} \\
c[w+u]=[c w+c u] \longmapsto[c w]=c[w]
\end{gathered}
$$

which shows the conditions for the mapping to be a $k$-linear morphism. Now notice that the mapping is clearly surjective, since given any class $[w] \in W /(W \cap U)$ we have the element $[w+u] \in(W+U) / U$, where $u$ is any element of $U$ and moreover $[w+u] \mapsto[w]$ from construction, showing the surjective character of the mapping. Moreover, notice that $[w] \in W /(W \cap U) \Leftrightarrow w \in U$, thus the kernel of the mapping is equal to $U$. Now by means of the first isomorphism theorem shows that the mapping is also injective.

Theorem 5.4.6 (Third Isomorphism). Let $V$ be a $k$-vector space and let the subspaces $U \subseteq W \subseteq V$. Then there exists an embedding $W / U \hookrightarrow V / U$ with a mapping [ $w$ ] $\mapsto$ [ $w$ ] such that $W / U$ can be regarded as a subspace of $V / U$. Moreover, we have a natural isomorphism

$$
V / W \simeq(V / U) /(W / U)
$$

Proof. Consider the morphism $\varphi: V / U \rightarrow V / W$ with the mapping $[v] \stackrel{\varphi}{\mapsto}[v]$. Since $U \subseteq W$ then we have that $[v]=\left[v^{\prime}\right] \in V / U$ implies $[v]=\left[v^{\prime}\right] \in V / W$, which shows that $\varphi$ is well defined. Moreover, given a class $[v] \in V / W$ we have the corresponding class $[v] \in V / U$ such that $[v] \stackrel{\varphi}{\mapsto}[v]$, thus $\varphi$ is surjective.

Moreover, notice that if $[w] \in W / U$ then from the embedding we have the corresponding class $[w] \in V / U$, we'll have two possibilities, if $[w]=[0]$, then the $k$-linear morphism $\varphi$ will map it to [0], on the other hand, if $[w] \neq[0]$ then $w \in W$ implies that $[w] \stackrel{\varphi}{\mapsto}[w]=[0] \in V / W$ so that in both cases we have $[w] \in W / U \Rightarrow[w] \in \operatorname{ker} \varphi$ so that $W / U \subseteq \operatorname{ker} \varphi$. Moreover, if $V / U \ni[v] \in \operatorname{ker} \varphi$, then $[v]=[0] \in V / W$ which means that $v \in W$ and therefore we have the corresponding class $[v] \in W / U$, which sums up to $\operatorname{ker} \varphi \subseteq W / U$ and thus $\operatorname{ker} \varphi=W / U$. From the first isomorphism theorem we have that

$$
(V / U) /(W / U) \simeq V / W
$$

since the image of $\varphi$ is equal to $V / W$, settling the proof.

### 5.5 Vector Spaces from Linear Morphism

Proposition 5.5.1. Given $k$-vector spaces $V$ and $W$, consider the $k$-linear morphism $\varphi: V \rightarrow W$. Then $\varphi$ is injective if and only if $\operatorname{ker} \varphi=\{0\}$.

Proof. $(\Rightarrow)$ Suppose $\varphi$ is injective, then, since $0 \stackrel{\varphi}{\mapsto} 0$, it follows that $\operatorname{ker} \varphi=\{0\}$. $(\Leftarrow)$ Suppose now that $\operatorname{ker} \varphi=\{0\}$, then from the universal property of quotients we find that $V / \operatorname{ker} \varphi \mapsto W$ is injective, since the only possible element in $V$ that generates the zero class is $0 \in V$, therefore, we can regard the above mapping as [v] $\mapsto \varphi(v)$, therefore we conclude that $\varphi$ is injective.

Theorem 5.5.2 (Universal property for kernels). Let $V, W$ be $k$-vector spaces and consider the $k$-linear morphism $\varphi: V \rightarrow W$. Denote $t: \operatorname{ker} \varphi \hookrightarrow V$ the embedding morphism. Then, for any given $k$-vector space $L$ and $k$-linear morphism $\psi: L \rightarrow V$ such that $\varphi \psi=0$, there exists a unique $k$-linear morphism $\ell: L \rightarrow \operatorname{ker} \varphi$ such that the following diagram commutes


Proof. The condition for the commutativity of the diagram is $\psi=\iota \ell$. Notice that since the definition of $\psi$ depends uniquely in the condition $\varphi \psi=0$, we find that since $\operatorname{im}(\ell) \subseteq \operatorname{ker} \varphi$ it follows that indeed $\varphi(\iota \ell)=0$ and therefore the uniqueness of $\ell$ comes by defining it as $\ell(l)=\psi(l)$ and the existence comes merely by the fact that $\psi$ is already a $k$-linear morphism.

Definition 5.5.3 (Cokernel). Let $k$-vector spaces $V$ and $W$ and consider the linear morphism $\varphi: V \rightarrow W$. We define the cokernel of $\varphi$ as

$$
\operatorname{coker} \varphi=W / \operatorname{im}(\varphi)
$$

Theorem 5.5.4 (Universal property for cokernels). Let $k$-vector spaces $V$ and $W$ and the linear morphism $\varphi: V \rightarrow W$. Denote by $\pi: W \rightarrow \operatorname{coker} \varphi$ the projection morphism. For any given $k$-vector space $L$ and linear morphism $\psi: W \rightarrow L$ for which $\psi \varphi=0$, there exists a unique linear morphism $\ell: \operatorname{coker} \varphi \rightarrow L$ such that the diagram commutes


Proof. The condition for the diagram to commute is that $\ell \pi=\psi$. For that, notice that since coker $\varphi=W / \operatorname{im}(\varphi)$ it follows that $\ell \pi \varphi(v)=\ell([\varphi(v)])$ but notice that obviously $\varphi(v) \in \operatorname{im}(\varphi)$ thus $[\varphi(v)]=0 \in \operatorname{coker} \varphi$ thus indeed $\ell \pi \varphi=0$. Moreover, in order to define $\ell$ we ought to have $\ell([w])=\psi(w)$, which is well defined since $[w]=w+\operatorname{im}(\varphi)$ and $\forall w^{\prime} \in \operatorname{im}(\varphi)$ we have $\psi\left(w^{\prime}\right)=0$ and therefore $\psi\left(w+w^{\prime}\right)=\psi(w)$, which ends the proof since $\ell$ gets the linear structure of $\psi$ and also is unique from construction. $\quad$.

Proposition 5.5.5. Let $k$-vector spaces $V$ and $W$ and define any linear morphism $\varphi: V \rightarrow$ $W$, then there are natural isomorphisms

$$
\operatorname{ker}(W \xrightarrow{\pi} \operatorname{coker} \varphi) \simeq \operatorname{coker}(\operatorname{ker} \varphi \xrightarrow{\iota} V) \simeq \operatorname{im}(\varphi)
$$

Proof. First we show the existence of the isomorphism $\operatorname{ker} \pi \simeq \operatorname{im}(\varphi)$, which is obtained by taking the mapping $w \mapsto w$, since $\operatorname{ker} \pi=\operatorname{im}(\pi)$, establishing an obvious isomorphism between the two given objects. Now we focus on showing the isomorphism coker $\iota=V / \operatorname{ker} \varphi \simeq \operatorname{im}(\varphi)$, which can be obtained by considering the mapping $[v] \mapsto \varphi(v)$, which is well defined from the same reasoning as in the above theorem. $\quad \square$

Consider the collections of $k$-vector spaces $\left\{V_{i}\right\}_{i}$ and $\left\{W_{i}\right\}_{i}$ with a corresponding collection of morphism $\left\{f_{i}: V_{i} \rightarrow W_{i}\right\}_{i}$ and $\left\{\varphi_{i}: V_{i} \rightarrow V_{i+1}\right\}_{i}$ and $\left\{\psi_{i}: W_{i} \rightarrow W_{i+1}\right\}$ such that for any index $i$ the following diagram commutes


By defining the morphisms $\iota_{i}: \operatorname{ker} f_{i} \rightarrow V$ and $\pi_{i}: W \rightarrow \operatorname{coker} f_{i}$ we find that since

$$
f_{i+1} \varphi_{i} l_{i}=\psi_{i} f_{i} l_{i}=\psi_{i} 0=0
$$

then we can use the universal property of kernels in order to find the existence of a unique morphism $\operatorname{ker} f_{i} \rightarrow \operatorname{ker} f_{i+1}$, that is


On the other hand we have that

$$
\pi_{i+1} \psi_{i} f_{i}=\pi_{i+1} f_{i+1} \varphi_{i}=0 \varphi_{i}=0
$$

thus by means of the universal property for cokernels we find that there exists a unique morphism coker $f_{i} \rightarrow$ coker $f_{i+1}$, that is


Binding both results together we find that in general we have the following diagram


Definition 5.5.6 (Exact sequence). Given a collection of $k$-vector spaces $\left\{V_{i}\right\}_{i}$ and $k$ linear morphisms $\left\{f_{i}: V_{i} \rightarrow V_{i+1}\right\}$, we say that the sequence

$$
V_{0} \xrightarrow{f_{0}} V_{1} \xrightarrow{f_{1}} \ldots \xrightarrow{f_{n-1}} V_{n}
$$

is exact if $\operatorname{ker} f_{i}=\operatorname{im}\left(f_{i-1}\right)$ for all index.
Proposition 5.5.7. Let the $k$-linear morphism $f: V \rightarrow W$. Then $f$ is injective if and only if the sequence $0 \longrightarrow V \xrightarrow{f} W$ is exact. On the other hand, the morphism $f$ is surjective if and only if the sequence $V \xrightarrow{f} W \longrightarrow 0$ is exact.

Proof. $(\Rightarrow)$ Suppose $f$ is injective, then $\operatorname{ker} f=\{0\}$, moreover, surely $\operatorname{im}(0)=\operatorname{ker} f$ thus the sequence is exact. $(\Leftarrow)$ Suppose the sequence is exact, so that $\operatorname{im}(0)=\operatorname{ker} f$ but then $\operatorname{ker} f=\{0\}$ which, as we already have proven above, implies in $f$ injective.
$(\Rightarrow)$ Suppose $f$ is surjective, then $\operatorname{im}(f)=W=\operatorname{ker} 0$ and thus the sequence is exact. $(\Leftarrow)$ Suppose the sequence is exact, so that $\operatorname{im}(f)=W=$ ker 0 , then given any element $w \in W$ there exists at least one corresponding $v \in V$ such that $f(v)=w$ and therefore the morphism is surjective.

Definition 5.5.8 (Short exact sequence). Let the $k$-vector spaces $V, W$ and $L$. A exact sequence

$$
0 \longrightarrow W \xrightarrow{f} V \xrightarrow{g} L \longrightarrow 0
$$

that is, with $\operatorname{ker} g=\operatorname{im}(f)$, is said to be a short exact sequence of $k$-vector spaces.
Proposition 5.5.9. In the short exact sequence $0 \rightarrow W \stackrel{f}{\rightarrow} V \xrightarrow{g} L \rightarrow 0$ we have $L \simeq$ V/W.

Proof. Firstly, notice that since $W \mapsto V$ is injective, we can regard $W$ as a subspace of $V$, so that taking the quotient $V / W$ is possible. Moreover, notice that since $g$ is surjective, from the first isomorphism theorem we find that $V / \operatorname{ker} g \simeq L$ which from the fact that the sequence is exact, is the same as $V / \operatorname{im} f \simeq L$ but this is exactly what we meant by $V / W$ since $W$ was regarded as a subspace of $V$ via the embedding $f$.

Proposition 5.5.10. Let $f: V \rightarrow W$ be a $k$-linear morphism, then the following is true
(a). The sequence $0 \rightarrow \operatorname{ker} f \rightarrow V \rightarrow \operatorname{im} f \rightarrow 0$ is exact.
(b). The sequence $0 \rightarrow \operatorname{im} f \rightarrow W \rightarrow$ coker $f \rightarrow 0$ is exact.
(c). The sequence $0 \rightarrow \operatorname{ker} f \rightarrow V \rightarrow W \rightarrow \operatorname{coker} f \rightarrow 0$ is exact.

Proof. For (I) notice that trivially $\operatorname{ker}(\operatorname{ker} f \rightarrow V)=\operatorname{im}(0)=\{0\}$, since $\operatorname{ker} f \subseteq V$; moreover, $\operatorname{im}(\operatorname{ker} f \rightarrow V)=\operatorname{ker} f=\operatorname{ker}(V \rightarrow \operatorname{im}(f))$ from the mere definition of kernel; finally, it is to be noticed that $\operatorname{im}(V \rightarrow \operatorname{im}(f))=\operatorname{ker}(\operatorname{im}(f) \rightarrow 0)$ because the zero morphism has its whole domain as its kernel. For (II) we have that $\operatorname{im}(0 \rightarrow$ $\operatorname{im}(f))=\{0\} \subseteq W$, thus indeed $\operatorname{ker}(\operatorname{im}(f) \rightarrow W)=\operatorname{im}(0 \rightarrow \operatorname{im}(f))$; after that, notice $\operatorname{im}(\operatorname{im}(f) \rightarrow W)=\operatorname{im}(f)$ but from definition we have coker $f=W / \operatorname{im}(f)$ thus we really get $\operatorname{im}(f)=\operatorname{ker}($ coker $f)$; the next morphism trivially obeys $\operatorname{im}(W \rightarrow \operatorname{coker} f)=$ $\operatorname{ker}(\operatorname{coker} f \rightarrow 0)$. For (III) the part $0 \rightarrow \operatorname{ker} f \rightarrow V$ is already exact, also from definition $\operatorname{im}(\operatorname{ker} f \rightarrow V)=\operatorname{ker} f=\operatorname{ker}(V \rightarrow W)$; then, $\operatorname{im}(V \rightarrow W)=\operatorname{im}(f)=\operatorname{ker}($ coker $f)$; and finally $\operatorname{im}(W \rightarrow \operatorname{coker} f)=\operatorname{ker}(\operatorname{coker} f \rightarrow 0)$.

Definition 5.5.11 (The Mor space). Let the $k$-vector spaces $V, W$. Then the collection of morphisms $\operatorname{Mor}_{\text {Vect }_{k}}(V, W)$ is a $k$-vector space with structure given by

$$
\begin{gathered}
(f+g)(v)=f(v)+g(v), \forall f, g \in \operatorname{Mor}_{\text {vect }_{k}}(V, W) \\
(c f)(v)=c f(v), \forall f \in \operatorname{Mor}_{\text {Vect }_{k}}(V, W) \forall c \in k
\end{gathered}
$$

where 0 is the map $0: V \rightarrow 0 \rightarrow W$.

Remark 5.5.12. For the remaining this section we'll be dealing with the category Vect ${ }_{k}$, unless said otherwise, therefore I'll omit the category for the sake of notation.

Proposition 5.5.13. For any $k$-vector space $V$ there exists a natural isomorphism

$$
\operatorname{Mor}(k, V) \simeq V
$$

Proof. Map the zero morphism $(0: k \rightarrow V) \longmapsto 0 \in V$. Then we could take the morphism $1_{v}: k \rightarrow V$ being defined as $k \ni 1 \stackrel{1_{v}}{\longmapsto} v \in V$, not restricting the other mappings whatsoever, then we could trivially make the mappings $\left(1_{v}: k \rightarrow V\right) \longmapsto v \in$ $V$ and then we are essentially done, since this is trivially an isomorphism.

Proposition 5.5.14. Let the collection of $k$-vector spaces $\left\{V_{i}\right\}_{i \in I}$, then, for any $L, k$-vector space, then:
I. There is a natural isomorphism

$$
\operatorname{Mor}\left(L, \prod_{i \in I} V_{i}\right) \simeq \prod_{i \in I} \operatorname{Mor}\left(L, V_{i}\right)
$$

II. There is a natural isomorphism

$$
\operatorname{Mor}\left(\bigoplus_{i \in I} V_{i}, L\right) \simeq \prod_{i \in I} \operatorname{Mor}\left(V_{i}, L\right)
$$

and therefore, given a set $S$, there is also a natural isomorphism $\operatorname{Mor}\left(k^{\oplus S}, V\right) \simeq V^{S}$.
Proof. For the first proposition, we can assign the map of zero mappings ( $0: L \rightarrow$ $\left.\Pi V_{i}\right) \longmapsto\left(0: L \rightarrow V_{i}\right)_{i}$. Next we may consider the functions $g: L \rightarrow \Pi V_{i}$, defined as $g(l)=\left(f_{1}(l), f_{2}(l), \ldots\right)$ for all $l \in L$, where $f_{i}: L \rightarrow V_{i}$; then we can simply make the mapping $g \longmapsto\left(\pi_{i} f_{i}\right)_{i}$.

For the second proposition, consider the morphism $\psi$ : $\operatorname{Mor}\left(\bigoplus_{i \in I} V_{i}, L\right)$ with the mapping $f \mapsto\left(f \iota_{i}\right)_{i \in I}$ where $\iota_{j}: V_{j} \hookrightarrow \bigoplus_{i \in I} V_{i}$ is the inclusion map. For the injectivity, consider $f \in \operatorname{Mor}\left(\bigoplus_{i \in I} V_{i}, L\right), f \neq 0$ be a morphism, then $f \iota_{i} \neq 0$ and therefore $\operatorname{ker}(\psi)=0$, which implies in the injectivity of $\psi$. Now, let any tuple of morphisms $\left(g_{i}\right)_{i \in I} \in \prod_{i \in I} \operatorname{Mor}\left(V_{i}, L\right)$, then we can define a morphism $f \in \operatorname{Mor}\left(\bigoplus_{i \in I} V_{i}, L\right)$ such that $f \iota_{i}=g_{i}$ then definitely $\psi(f)=\left(g_{i}\right)_{i \in I}$.

Definition 5.5.15 (Induced Mor $k$-linear morphism). Given a $k$-linear morphism $f: V \rightarrow$ $L$ and any $k$-vector space $W$, there are induced, uniquely defined, $k$-linear morphisms
(a) (Pushforward) $f_{*}: \operatorname{Mor}(W, V) \rightarrow \operatorname{Mor}(W, L)$ with the mapping $\alpha \mapsto f \alpha$, so that the diagram commutes

(b) (Pullback) $f^{*}: \operatorname{Mor}(L, W) \longrightarrow \operatorname{Mor}(V, W)$ with the mapping $\alpha \longmapsto \alpha f$, so that the diagram commutes


Proposition 5.5.16. Given $V_{1}, V_{2}, V_{3}, k$-vector spaces, the following holds for any given $k$-vector space $L$ :
I. If the sequence $0 \rightarrow V_{1} \rightarrow V_{2} \rightarrow V_{3}$ is exact, then

$$
\operatorname{Mor}(L, 0)=0 \rightarrow \operatorname{Mor}\left(L, V_{1}\right) \rightarrow \operatorname{Mor}\left(L, V_{2}\right) \rightarrow \operatorname{Mor}\left(L, V_{3}\right)
$$

is an exact sequence, that is, covariant Mor is left exact.
II. If the sequence $V_{1} \rightarrow V_{2} \rightarrow V_{3} \rightarrow 0$ is exact, then

$$
\operatorname{Mor}(0, L)=0 \rightarrow \operatorname{Mor}\left(V_{3}, L\right) \rightarrow \operatorname{Mor}\left(V_{2}, L\right) \rightarrow \operatorname{Mor}\left(V_{1}, L\right)
$$

is an exact sequence, that is, contravariant Mor is left exact.
Proof. For each of the propositions, we'll denote the morphisms $f: V_{1} \rightarrow V_{2}$ and $g: V_{2} \rightarrow$ $V_{3}$ such that $\operatorname{ker} g=\operatorname{im} f$.

For the first proposition, let $0 \rightarrow V_{1} \rightarrow V_{2} \rightarrow V_{3}$ be a an exact sequence, and consider the sequence $0 \rightarrow \operatorname{Mor}\left(L, V_{1}\right) \xrightarrow{f_{*}} \operatorname{Mor}\left(L, V_{2}\right) \xrightarrow{g_{*}} \operatorname{Mor}\left(L, V_{3}\right)$. Let $\beta \in \operatorname{im} f_{*}$, then there exists $\alpha \in \operatorname{Mor}\left(L, V_{1}\right)$ such that $f_{*}(\alpha)=f \alpha=\beta$, hence $g_{*}(\beta)=g \beta=g f \alpha=0$ since $\operatorname{ker} g=\operatorname{im} f$, which implies in $\operatorname{im} f_{*} \subseteq \operatorname{ker} g_{*}$. Suppose now that $\beta \in \operatorname{ker} g_{*}$, so that $g_{*}(\beta)=g \beta=0$, then we find that $\operatorname{im} \beta \subseteq \operatorname{ker} g=\operatorname{im} f$ and hence $\beta \in \operatorname{im} f_{*}$, which implies in $\operatorname{ker} g_{*} \subseteq \operatorname{im} f_{*}$. Therefore $\operatorname{ker} g_{*}=\operatorname{im} f_{*}$.

For the second proposition, let $V_{1} \rightarrow V_{2} \rightarrow V_{3} \rightarrow 0$ be a an exact sequence and consider the sequence $0 \rightarrow \operatorname{Mor}\left(V_{3}, L\right) \xrightarrow{g^{*}} \operatorname{Mor}\left(V_{2}, L\right) \xrightarrow{f^{*}} \operatorname{Mor}\left(V_{1}, L\right)$. Let $\gamma \in \operatorname{im} g_{*}$ and consider $\lambda \in \operatorname{Mor}\left(V_{3}, L\right)$ such that $g^{*}(\lambda)=\lambda g=\gamma$. Then we find that $f^{*}(\gamma)=\gamma f=$ $\lambda g f=\lambda 0=0$ hence $\gamma \in \operatorname{ker} f^{*}$, which implies in im $g^{*} \subseteq \operatorname{ker} f^{*}$.

Let $\gamma \in \operatorname{ker} f^{*}$ so that $f^{*}(\gamma)=\gamma f=0$, this implies in $\operatorname{im} f=\operatorname{ker} g \subseteq \operatorname{ker} \gamma$. From Theorem 5.4.3 we find that $\gamma$ induces a morphism $\bar{\gamma}$ such that the following diagram commutes

so that $\gamma=\bar{\gamma} \pi$. Moreover, since $g$ is surjective we have $\operatorname{im} g=V_{3}$, and from Theorem 5.4.4 we find that $g$ induces an isomorphism $\bar{g}: V_{2} / \operatorname{ker} g \xrightarrow{\simeq} V_{3}$, so that the
following diagram commutes

so that $g=\bar{g} \pi$ and since $\bar{g}$ is an isomorphism, then $\pi=g \bar{g}^{-1}$. Notice now that $g^{*}\left(\overline{\gamma g}^{-1}\right)=\overline{\gamma g}^{-1} g=\bar{\gamma} \pi=\gamma$ and therefore $\gamma \in \operatorname{im} g^{*}$. This shows that ker $f^{*} \subseteq \operatorname{im} g^{*}$ and therefore $\operatorname{ker} f^{*}=\operatorname{im} g^{*}$.

### 5.6 Bases and Dimensions

Definition 5.6.1. Let $V$ a $k$-vector space and $S \subseteq V$ a subset. By the universal property of free vector spaces, the inclusion $S \hookrightarrow V$ induces the unique $k$-linear morphism $f: k^{\oplus S} \rightarrow V$. Then we can classify the set $S$ in terms of $f$ as:
I. Linearly independent, in the case where $f$ is injective, and linearly dependent otherwise.
II. Generating (or spanning) if $f$ is surjective.
III. Basis if $f$ is an isomorphism.

Proposition 5.6.2. A set $S \subseteq V$ is linearly independent if and only if for all subset $\left\{v_{i}\right\}_{i} \subseteq S$ and scalars $\left\{a_{i}\right\}_{i} \subseteq k$, the equation $\sum_{i} a_{i} v_{i}=0$ implies $a_{i}=0$ for all index $i$.

Definition 5.6.3 (Subspace spanned). Let $V$ be a $k$-vector space. Given a set $S \subseteq V$, we define the subspace spanned by $S$ to be

$$
\operatorname{span}(S)=\operatorname{im}\left(k^{\oplus S} \rightarrow V\right)
$$

Proposition 5.6.4. Let $V$ be a $k$-vector space and the set $S \subseteq V$. Then $S$ is a basis for $V$ if and only if for all $v \in V$ there exists a unique tuple $\left(a_{s}\right)_{s \in S} \subseteq k$ such that there are only finitely many $a_{s} \neq 0$ and $\sum_{s \in S} a_{s} s=v$.

Lemma 5.6.5 (Zorn's Lemma). Let $X$ be a set and consider the partial order relation $(\vdash) \subseteq X \times X$, so that $\vdash$ is reflexive for all $x \in X$, antisymmetric for all $x, y \in X$, and transitive for all $x, y, z \in X$. Moreover, if $S \subseteq X$ is such that only one of the propositions: $x \vdash y$ or $y \vdash x$ is true for every $x, y \in S$; then there exists an element $z \in X$ for which $x \vdash z$ for all choices of $x \in X$. Then there exists an element $m \in X$ for which $m \vdash x$ if and only if $x=m$.

Theorem 5.6.6. Every vector space has a basis.

Proof. Let $V$ be a $k$-vector space. Consider the set $\mathcal{A}:=\{X \subseteq V: X$ linearly independent $\}$ partially order under inclusion, and let $C \subseteq \mathcal{A}$ be a totally ordered subset by inclusion.

We now show that $\mathcal{B}$ satisfies the hypothesis needed for the use of Zorn's lemma. Define $Z$ to be the union of all sets in $C$-we'll show that $Z$ is linearly independent. For the sake of contradiction, suppose $Z$ is linearly dependent, then one must be able to pick a finite collection $\left(v_{j}\right)_{j=1}^{n}$ of members of $Z$ such that there exists an associated collection $\left(a_{j}\right)_{j=1}^{n}$ of elements $a_{j} \in k$, not all zero, such that

$$
\begin{equation*}
a_{1} v_{1}+\cdots+a_{n} v_{n}=0 \tag{5.1}
\end{equation*}
$$

Since each $v_{j} \in S_{j}$ for some set $S_{j} \in C$, then by the total ordering of $C$ we may assume that

$$
S_{1} \subseteq S_{2} \subseteq \cdots \subseteq S_{n-1} \subseteq S_{n}
$$

up to change of indexing. Since $v_{j} \in S_{n}$ for all $1 \leqslant j \leqslant n$, then in particular Eq. (5.1) shows that $S_{n}$ is linearly dependent, which is false by hypothesis that $S_{n} \in \mathcal{A}$. We conclude that $Z$ must be linearly independent and thus $Z \in \mathcal{A}$.

Since $\mathcal{A}$ is ordered by inclusion, we can apply Zorn's lemma to conclude that $\mathcal{A}$ admits a maximal element $M \in \mathcal{A}$. We'll show that $M$ does generate $V$. Suppose, for the sake of contradiction, that there exists $v \in V$ such that $v$ is not generated by $M$. Then it must be the case that $M \cup\{v\}$ is linearly-independent hence $M \cup\{v\} \in \mathcal{A}$ and $M \subseteq M \cup\{v\}$, which is a contradiction since $M$ is maximal. Therefore $\operatorname{span}(M)=V$, which shows ${ }^{1}$ that $M$ is a basis for $V$.

Corollary 5.6.7. Every vector space is isomorphic to $k^{\oplus S}$ for some set $S$.
Proposition 5.6.8. Let $B$ be a basis of the $k$-vector space $V$. Then for every $k$-vector space $W$ there is a natural isomorphism of vector spaces

$$
\operatorname{Mor}_{\operatorname{Vect}_{k}}(V, W) \simeq \prod_{b \in B} W
$$

Proof. We show that the map

$$
(f: V \rightarrow W) \longmapsto(f(b))_{b \in B}
$$

is an isomorphism. First, it is injective: let $f, g \in \operatorname{Mor}(V, W)$; suppose that $f=g$ so that obviously $(f(b))_{b \in B}=(g(b))_{b \in B}$ since $f(b)=g(b)$ for all $b \in B$; suppose now that $(f(b))_{b \in B}=(g(b))_{b \in B}$, since $B$ is a basis on $V$ then, given any $v \in V$, we can write it as $v=\sum_{b \in B} a_{b} b$ where $a_{b} \neq 0$ only finitely many and $a_{b} \in k$, then

$$
f(v)=f\left(\sum_{b \in B} a_{b} b\right)=\sum_{b \in B} a_{b} f(b)=\sum_{b \in B} a_{b} g(b)=g\left(\sum_{b \in B} a_{b} b\right)=g(v)
$$

thus indeed the mapping is injective. Now we show that it is surjective: notice that $|W|^{|B|}=\left|\prod_{b \in B} W\right|$ and since the morphism of $\operatorname{Mor}(V, W)$ are completely determined by its image under $B \subseteq V$ then we conclude that $|\operatorname{Mor}(V, W)|=|W|^{|B|}$, thus we can create a mapping which satisfies the surjectivity.

[^7]Proposition 5.6.9. Let $\alpha: S \rightarrow T$ be a map between sets. Then the induced map $\alpha_{*}: k^{\oplus S} \rightarrow k^{\oplus T}$ is such that
i. $\alpha_{*}$ is an monomorphism if $\alpha$ is injective.
ii. $\alpha_{*}$ is an epimorphism if $\alpha$ is surjective.
iii. $\alpha_{*}$ is an isomorphism if $\alpha$ is bijective.

Proof. For the first, let $\alpha$ be injective, then given $f \in k^{\oplus S}$ and define the finite set $S_{f}:=S \backslash \operatorname{ker} f$. Now we construct a unique function $g \in k^{\oplus T}$ defined such that $g(t) \neq 0$ if and only if $t \in \alpha\left(S_{f}\right)$ (well defined since $\alpha\left(S_{f}\right)$ is finite by the injectivity of $\alpha$ ), and moreover the condition $f(s)=g(\alpha(s))$ (uniqueness), then our mapping $\alpha *$ can be defined as $f \mapsto g$, which is surely a monomorphism and well defined.

For the second proposition, let $\alpha$ be surjective, then from the same map as before, given any function $g \in k^{\oplus T}$, let $T_{g}:=T \backslash \operatorname{ker} g$ then for every $t \in T_{g}$ we chose one $s \in \alpha^{-1}(t)$ so that for some function $f \in k^{\oplus T}$ defined by $f(s)=g(\alpha(s))=g(t)$ and moreover $f(s) \neq 0$ if and only if $s \in \alpha^{-1}(t)$ for some $t \in T_{g}$ (thus $f$ is well defined, since $T_{g}$ is finite and we are taking only one corresponding $s$ for each $t \in T_{g}$ ). Thus indeed $f \stackrel{\alpha_{*}}{\longmapsto} g$ is surjective.

The last proposition comes trivially from the last two.
Proposition 5.6.10. Let the surjective $k$-linear morphism $f: V \rightarrow W$. There exists an injective $k$-linear morphism $g: W \mapsto V$ such that $f g=\mathrm{id}_{W}$.

Proof. Just take the mapping $w \stackrel{g}{\longmapsto} v \in f^{-1}(w) \in V$, then $w \stackrel{g}{\mapsto} v \stackrel{f}{\mapsto} w$ and thus $f g=\mathrm{id}_{W}$.

Proposition 5.6.11. Let the injective $k$-linear morphism $f: V \mapsto W$. There exists a surjective morphism $g: W \rightarrow V$ such that $g f=\mathrm{id}_{V}$.

Proof. Given any $v \in V$, we want $g(f(v))=g(w)=v$, but since $f^{-1}(w)$ is a singleton, we can simply make the well defined mapping $w \stackrel{g}{\mapsto} v$ for all $w \in \operatorname{im}(f)$. With this unique condition we already have the wanted $g f=\mathrm{id}_{V}$. Moreover, given any $v \in V$, there must exist $w \in W$ (in fact we could specify $w \in \operatorname{im}(f)$, but here we can be more general) such that $g(w)=v$, thus $g$ is surjective.

Proposition 5.6.12. Let $k$-vector spaces $V$ and $W$. There exists injective linear morphism $V \mapsto W$ if and only if there exists a surjective morphism $W \rightarrow V$.

Proof. The last two propositions.
Theorem 5.6.13. Let sets $S$ and $T$. The following propositions are equivalent:
I. There exists injective linear morphism $k^{\oplus S} \mapsto k^{\oplus T}$.
II. There exists surjective linear morphism $k^{\oplus T} \rightarrow k^{\oplus S}$.
III. There exists an injection $S \mapsto T$.
IV. There exists a surjective map $T \rightarrow S$ or $S=\varnothing$.

Lemma 5.6.14. Let $V$ a $k$-vector space, $B \subseteq V$ a basis and $S \subseteq V$ a spanning set. Then for all $b \in B$ there exists $a \in A$ such that $(B \backslash\{b\}) \cup\{a\}$ is a basis.

Proof. Let $b \in B$ be any element. Suppose, for the sake of contradiction, that $A \subseteq$ $\operatorname{span}(B \backslash\{b\})$ then $V=\operatorname{span}(A) \subseteq \operatorname{span}(B \backslash\{b\})$, which is a contradiction because $B$ is said to be a basis for $V$; thus $A \nsubseteq \operatorname{span}(B \backslash\{b\})$. Let $a \in A \backslash \operatorname{span}(B \backslash\{b\})$, then consider the set $(B \backslash\{b\}) \backslash\{a\}$. Let then $a=c b+c_{1} b_{1}+\cdots+c_{n} b_{n}$ with $c \neq 0$ from the construction of $a$ (that is, $a \notin \operatorname{span}(B \backslash\{b\})$ ). This way we can make

$$
b=a / c-c_{1} / c b_{1}-\cdots-c_{n} / c b_{n} \text { therefore } b \in \operatorname{span}((B \backslash\{b\}) \cup\{a\})
$$

which makes $(B \backslash\{b\}) \cup\{a\}$ a basis for $V$.
Lemma 5.6.15. Let $B$ and $B^{\prime}$ be bases the $k$-vector space $V$, then there exists a bijection $B \simeq B$, and thus $|B|=\left|B^{\prime}\right|$.

Proof. From Theorem 5.6.13 we see that since exists injective $k$-linear morphism $V \mapsto V$, then exists injective map $B \rightarrow B^{\prime}$ and also surjective map $B \rightarrow B^{\prime}$. From Cantor-Schröder-Bernstein theorem we see that there exists bijection $B \simeq B^{\prime}$.

Proposition 5.6.16. Let $V$ be a $k$-vector space and $S \subseteq V$ be any linearly independent set. There exists a set $B \supseteq S$ such that $B$ is a basis of $V$.

Proof. Let $\mathcal{S}$ be the non-empty collection of linearly independent sets of $V$ containing $S$. Notice that a the union of a chain of elements of $\mathcal{S}$ is again a linearly independent set containing $S$ - thus $\mathcal{S}$ is closed under arbitrary unions. In other words, every chain of elements has an upper bound in $\mathcal{S}$. By Zorn's lemma, it follows that the collection $\mathcal{S}$ has a maximal element - call it $B$. We now prove that $B$ is a basis for $V$.

Let $v \in V$ be an element such that $v \notin B$. Since $B$ is maximal, it follows that the set $B \cup\{v\}$ must not be linearly independent - since it contains $B$. This implies in the existence of a collection $\left(c_{b}\right)_{b \in B}$ of finitely many non-zero elements $c_{b} \in k$, and $c_{0} \in k$ such that

$$
c_{0} v+\sum_{b \in B} c_{b} b=0
$$

with not all coefficients equal to zero. Since $v$ isn't a member of $B$, it follows that $c_{0} \neq 0$, otherwise $B$ wouldn't be linearly independent. It follows that

$$
v=\sum_{b \in B}-\frac{c_{b}}{c_{0}} b
$$

therefore $v$ belongs to the span of $B$ in $V$. Therefore $B$ indeed generates $V$, making it a basis.

Lemma 5.6.17. Let $V$ be a $k$-vector space. If $B \subseteq V$ is a minimal generating set for $V$, then $B$ is a basis of $V$.

Proof. We must show that $B$ is linearly independent. If, on the contrary, $B$ where to be linearly dependent, let $\left(c_{b}\right)_{b \in B}$ be a collection of finitely many non-zero scalars, not all zero, such that $\sum_{b \in B} c_{b} b=0$. Choose a non-zero scalar $c_{b^{\prime}}$ from this family, then

$$
b^{\prime}=\sum_{b \in B \backslash\left\{b^{\prime}\right\}}-\frac{c_{b}}{c_{b^{\prime}}} b,
$$

which implies that $b^{\prime}$ can be written as a linear combination of the remaining elements of $B$. This in particular implies that $B \backslash\left\{b^{\prime}\right\}$ is a generating set of $V$, which contradicts the hypothesis that $B$ is minimal.

## Matrices and changes of base

Definition 5.6.18 (Change of basis operator). Let $C: V \rightarrow V$ be a linear operator of a finite dimensional $k$-vector space, $V \simeq k^{n}$, defined by the mapping $[v]_{B_{1}} \stackrel{C}{\mapsto}[v]_{B_{2}}$, where $[v]_{B_{1}}$ and $[v]_{B_{2}}$ are the representations of the vector $v$ in the basis $B_{1}$ and $B_{2}$ of $V$. We define the matrix representation $C_{B_{1}, B_{2}}: k^{n} \rightarrow k^{n}$ of the linear operator $C$ to be the change of basis matrix. Then if $B_{1}=\left\{e_{i}\right\}_{i=1}^{n}$, and $B_{2}=\left\{e_{j}^{\prime}\right\}_{j=1}^{n}$, and $C_{B_{1}, B_{2}}=\left[c_{i, j}\right]$, then the coefficients $c_{i, j}$ of $C_{B_{1}, B_{2}}$ must be such that

$$
v=\sum_{i=1}^{n} a_{i} e_{i}=\sum_{j=1}^{n} b_{j} e_{j}^{\prime}=\sum_{j=1}^{n} b_{j}\left(\sum_{i=1}^{n} c_{i, j} e_{i}\right)=\sum_{i=1}^{n}\left(\sum_{j=1}^{n} c_{i, j} b_{j}\right) e_{i}
$$

Proposition 5.6.19 (Change of basis, linear morphism). Let $f: V \rightarrow W$ be a $k$-linear morphism of finite dimensional spaces, $V \simeq k^{n}$ and $W \simeq k^{m}$. Let $M_{B, S}: k^{n} \rightarrow k^{m}$ and $M_{B^{\prime}, S^{\prime}}$ be the matrix representation of $f$ in the basis $B$ and $S$, where $B$ is a basis of $V$ and $S$ a basis of $W$. Let now $B^{\prime}$ and $S^{\prime}$ be basis for $V$ and $S$. Then, the matrix representation of $f$ in the basis $B^{\prime}, S^{\prime}$, that is, $M_{B^{\prime}, S^{\prime}}: k^{n} \rightarrow k^{m}$, is given by the conjugation

$$
M_{B^{\prime}, S^{\prime}}=C_{S, S^{\prime}}^{-1} M_{B, S} C_{B, B^{\prime}}
$$

where $C_{S, S^{\prime}}$ and $C_{B, B^{\prime}}$ are the change of basis matrix representations defined in Definition 5.6.18.

Proof. Consider the function $C_{S, S^{\prime}}^{-1}, f C_{B, B^{\prime}}: V \rightarrow W$, then we have the mapping

$$
[v]_{B} \stackrel{C_{B, B^{\prime}}}{\longmapsto}[v]_{B^{\prime}} \stackrel{f}{\mapsto}\left[f\left([v]_{B^{\prime}}\right)\right]_{S} \stackrel{C_{S, S^{\prime}}}{\longmapsto}\left[f\left([v]_{B^{\prime}}\right)\right]_{S^{\prime}}
$$

hence the statement follows directly from Proposition 5.2.5.
Corollary 5.6.20. In particular, if $f: V \rightarrow V$ is a $k$-linear operator, and $B, B^{\prime}$ are basis of $V$, then given the matrix representation of $f$ in base $B$, namely, $M_{B}$, we have that the matrix representation of $f$ in base $B^{\prime}$ is given by

$$
M_{B^{\prime}}=C^{-1} M_{B} C
$$

where $C$ is the matrix representation of the change of basis operator from base $B$ to $B^{\prime}$.

## Dimension and Rank plus Nullity Theorem

Definition 5.6.21. Let $V$ be a $k$-vector space. We define the dimension of $V$ as

$$
\operatorname{dim}_{k}(V)=|B|
$$

for $B \subseteq V$ basis of $V$. In particular, if $\operatorname{dim}_{k}(V)<\infty$ we say that $V$ is a finite dimensional $k$-vector space.

Lemma 5.6.22. Let $V$ be a $k$-vector space and $W \subseteq V$ a $k$-subspace, then

$$
\operatorname{dim}_{k}(V)=\operatorname{dim}_{k}(W)+\operatorname{dim}_{k}(V / W)
$$

Proof. Let $B_{W}=\left\{w_{i}\right\}_{i \in I}$ be a basis for $W$ and $B_{V / W}=\left\{q_{j}\right\}_{j \in J}$ a basis for $V / W$. Then for all $j \in J$ there exists $v_{j} \in V$ such that $q_{j}=\left[v_{j}\right] \in V / W$. Define the set $A:=\left\{v_{j}: j \in J\right\}$, then the map $\varphi: J \rightarrow A$ with the mapping $j \mapsto v_{j}$ is surely surjective; moreover, if $j, t \in J$ and $j \neq t$, then $q_{j} \neq q_{t}$ since $B_{V / W}$ is a basis, hence $\left[v_{j}\right] \neq\left[v_{t}\right]$ and the map $\varphi$ is injective. Then $\varphi$ is a bijection and thus $|A|=|J|$. Now we show that $A$ is a linearly independent set: let $J_{t} \subseteq J$ be a finite subset of $J$ with $t$ elements and $\alpha_{1}, \ldots, \alpha_{t} \in k$ be such that $\sum_{\ell=1}^{t} \alpha_{\ell} v_{j_{\ell}}=0 \in V$ then $\sum_{\ell=1}^{t} \alpha_{\ell}\left[v_{j_{\ell}}\right]=\left[\sum_{\ell=1}^{t} \alpha_{\ell} v_{j_{\ell}}\right]=[0] \in V / W$ but notice that the set of classes $\left[v_{j}\right]: v_{j} \in A$ is linearly independent (from the fact that this corresponds to the basis $\left.B_{V / W}\right)$, hence we must have $\alpha_{\ell}=0$ for all $1 \leqslant \ell \leqslant t$, thus $A$ is linearly independent.

Consider now the set $A \cup B_{W}$, we'll show that this is a basis for $V$. First we show that $A \cup B_{W}$ is linearly independent: let $I_{s} \subseteq I$ be the finite set with $s$ indices, and $J_{t} \subseteq J$ be the finite set with $t$ indices, and $\alpha_{1}, \ldots \alpha_{s}, \beta_{1}, \ldots, \beta_{t} \in k$, be such that $\sum_{\ell=1}^{s} \alpha_{\ell} w_{i_{\ell}}+\sum_{r=1}^{t} \beta_{r} v_{j_{r}}=0 \in V$ if we now take the module with respect to $W$ (as we did before with $A$ ) we see that

$$
[0]=\sum_{\ell=1}^{s} \alpha_{\ell}\left[w_{i_{\ell}}\right]+\sum_{r=1}^{t} \beta_{r}\left[v_{j_{r}}\right]=\left[\sum_{r=1}^{t} \beta_{r} v_{j_{r}}\right] \in V / W
$$

since $w_{i_{\ell}} \in W$ and thus $\left[w_{i_{\ell}}\right]=[0] ;$ since $B_{V / W}$ (and $\left[v_{j_{r}}\right]=q_{j_{r}}$ ) is linearly independent, we see that necessarily $\beta_{r}=0$ for all $1 \leqslant r \leqslant t$. Therefore, we conclude that $\sum_{\ell=1}^{s} \alpha_{\ell} w_{i_{\ell}}=$ 0 and since $B_{W}$ is linearly independent, we conclude that $\alpha_{\ell}=0$ for all $1 \leqslant \ell \leqslant s$. For the final part, we need to show that $V=\operatorname{span}\left(A \cup B_{W}\right)$ : let $v \in V$ be any element, then $[v] \in V / W=\operatorname{span}\left(B_{V / W}\right)$ then $[v]=\sum_{\ell=1}^{t} \alpha_{\ell}\left[v_{j}\right]$ for $J_{t} \subseteq J$ and $\alpha_{1}, \ldots, \alpha_{t} \in k$; this implies that $[v]-\left[\sum_{\ell=1}^{t} \alpha_{\ell} v_{j_{\ell}}\right]=[0]$ and thus $v-\sum_{\ell=1}^{t} \alpha_{\ell} v_{j_{\ell}} \in W=\operatorname{span}\left(B_{W}\right)$; hence for some $\beta_{1}, \ldots, \beta_{s} \in k$ and $I_{s} \subseteq I$ we have $v-\sum_{\ell=1}^{t} \alpha_{\ell} v_{j_{\ell}}=\sum_{r=1}^{s} \beta_{r} w_{i_{r}}$ hence $v \in \operatorname{span}\left(A \cup B_{W}\right)$ and thus $A \cup B_{W}$ is a basis for $V$.

We'll now show that $A \cap B_{W}=\varnothing$ : suppose $u \in A \cap B_{W}$, then there exists $i \in I$ and $j \in J$ such that $u=w_{i}$ and $u=v_{j}$ - thus $q_{j}=\left[v_{j}\right]=\left[w_{i}\right]=[0] \in V / W$ which cannot be true since $B_{V / W}$ is a basis. Therefore $\left|A \cup B_{W}\right|=|A|+\left|B_{W}\right|=|J|+\left|B_{W}\right|=$ $\left|B_{V / W}\right|+\left|B_{W}\right|=\operatorname{dim}_{k}(V)$ as wanted.

Theorem 5.6.23 (Rank plus nullity). Let $L: V \rightarrow W$ be a $k$-linear morphism, then

$$
\operatorname{dim}_{k}(V)=\operatorname{dim}_{k}(\operatorname{ker}(L))+\operatorname{rank}(L) .
$$

Proof. Notice that from the first isomorphism theorem we have $V / \operatorname{ker}(L) \simeq \operatorname{im}(L)$ then it follows that $\operatorname{dim}_{k}(V / \operatorname{ker}(L))=\operatorname{dim}_{k}(\operatorname{im}(L))=\operatorname{rank}(L)$ from the fact that $\operatorname{ker}(L)$ is a $k$-subspace of $V$, we conclude from Lemma 5.6.22 that $\operatorname{dim}_{k}(V)=\operatorname{dim}_{k}(\operatorname{ker}(L))+$ $\operatorname{dim}_{k}(V / \operatorname{ker}(L))=\operatorname{dim}_{k}(\operatorname{ker}(L))+\operatorname{rank}(L)$.

Corollary 5.6.24. Let $V$ and $W$ be $k$-vector spaces and $L: V \rightarrow W$ be a $k$-linear morphism. Then we have
(a). If $L$ is injective, then $\operatorname{dim}_{k}(V) \leqslant \operatorname{dim}_{k}(W)$.
(b). If $L$ is surjective, then $\operatorname{dim}_{k}(V) \geqslant \operatorname{dim}_{k}(W)$.
(c). If $L$ is an isomorphism then $\operatorname{dim}_{k}(V)=\operatorname{dim}_{k}(W)$.

Proof. (a) If $L$ is injective, then $\operatorname{ker}(L)=0$ and thus $\operatorname{dim}_{k}(\operatorname{ker}(L))=0$ and from Theorem 5.6 .23 we have $\operatorname{dim}_{k}(V)=\operatorname{rank}(L)$ and since $\operatorname{rank}(L) \leqslant \operatorname{dim}_{k}(W)$ since $\operatorname{im}(L) \subseteq W$ then we conclude that the proposition is true. (b) Since $L$ is surjective, then $\operatorname{im}(L)=W$ and in particular $\operatorname{rank}(L)=\operatorname{dim}_{k}(W)$, thus Theorem 5.6.23 implies that $\operatorname{dim}_{k}(V)=\operatorname{dim}_{k}(\operatorname{ker}(L))+\operatorname{dim}_{k}(W) \geqslant \operatorname{dim}_{k}(W)$. (c) Already proved before, but also comes from the last two propositions.

Corollary 5.6.25. Let $L: V \rightarrow W$ be a $k$-linear morphism of finite vector spaces $V$ and $W$ such that $\operatorname{dim}_{k}(V)=\operatorname{dim}_{k}(W)$. Then the following propositions are equivalent:
(a). $L$ is an isomorphism.
(b). $L$ is injective.
(c). $L$ is surjective.

Proof. $(\mathrm{b} \Rightarrow \mathrm{c})$ Suppose $L$ injective, then from Theorem 5.6 .23 we have $\operatorname{dim}_{k}(V)=$ $\operatorname{rank}(L)$ since $\operatorname{dim}_{k}(\operatorname{ker}(L))=0$ then we have $\operatorname{dim}_{k}(W)=\operatorname{rank}(L)$ but then $\operatorname{im}(L)=W$ and therefore $L$ is a surjective linear map. $(c \Rightarrow a)$ Suppose that $L$ is surjective, from from Theorem 5.6.23 we have $\operatorname{dim}_{k}(V)=\operatorname{dim}_{k}(\operatorname{ker}(L))+\operatorname{rank}(L)=\operatorname{dim}_{k}(\operatorname{ker}(L))+\operatorname{dim}_{k}(W)$, but from hypothesis we have $\operatorname{dim}_{k}(V)=\operatorname{dim}_{k}(W)$, then, if the given vector spaces $V$ and $W$ are finite dimensional we have $\operatorname{dim}_{k}(\operatorname{ker}(L))=0$ and thus $\operatorname{ker}(L)$ is the zero space, which implies that $L$ is injective, hence an isomorphism.

### 5.7 Dual Vector Spaces

Definition 5.7.1. Let $V \in \mathrm{Obj}^{\left(\operatorname{Vect}_{k}\right)}$, then, we define the dual vector space of $V$ to be

$$
V^{*}=\operatorname{Mor}_{\text {vect }_{k}}(V, k)
$$

and $f \in V^{*}$ is called a linear functional on $V$.
Proposition 5.7.2. For all sets $S$ there exists a natural isomorphism between the dual free vector space generated by $S$ and the function space $S \rightarrow k$, that is

$$
\left(k^{\oplus S}\right)^{*} \simeq k^{S} .
$$

Proof. This result comes directly from the free vector space universal property, that is

so that we can assign for each function $f \in k^{S}$ the corresponding unique $k$-linear morphism $\ell \in\left(k^{\oplus S}\right)^{*}$.

Proposition 5.7.3. Let $V$ be finite dimensional, then there exists a non-canonical isomorphism

$$
V \simeq V^{*}
$$

Proof. The proof here is rather arbitrary, we'll just construct explicitly an isomorphism by defining an ordered basis $B=\left(v_{1}, \ldots, v_{n}\right)$ of $V$, but we could build other isomorphisms (this is why the proposition states "non-canonical"). Define the $k$-linear morphism $\varphi\left(v_{i}\right)=e_{i}$ where $e_{i}$ is the $i$-th standard basis on $k^{n}$, then $\varphi$ is an isomorphims $V \simeq k^{n}$. Moreover, from the standard inner product, we have that $k^{n} \simeq\left(k^{n}\right)^{*}$, thus we are done, since

$$
V \xrightarrow{\varphi} k^{n} \simeq\left(k^{n}\right)^{*} \xrightarrow{f \mapsto f \varphi} V^{*}
$$

thus indeed $V \simeq V^{*}$.
Proposition 5.7.4. If $V$ is finite-dimensional, then $V^{*}$ is also finite dimensional and

$$
\operatorname{dim}_{k}(V)=\operatorname{dim}_{k}\left(V^{*}\right)
$$

Proposition 5.7.5. Given $k$-vector spaces $V$ and $W$, there exists a natural isomorphism

$$
(V \oplus W)^{*} \simeq V^{*} \oplus W^{*}
$$

Proof. Consider the mapping $\Phi:(V \oplus W)^{*} \rightarrow V^{*} \oplus W^{*}$ defined by $\Phi\left(f^{*}\right)=\left(f_{V}^{*}, f_{W}^{*}\right)$, where $f_{V}^{*}=\left.f^{*}\right|_{V \oplus\{0\}}$ and $f_{W}^{*}=\left.f^{*}\right|_{\{0\} \oplus W}$. We now show that $\Phi$ is an isomorphism. (Injective) Suppose $g^{*} \in \operatorname{ker} \Phi$, then necessarily $\left.g^{*}\right|_{V \oplus\{0\}}=0$ and $\left.g^{*}\right|_{\{0\} \oplus W}=0$, but since $g^{*}=\left.g^{*}\right|_{V \oplus\{0\}}+\left.g^{*}\right|_{\{0\} \oplus W}$ we conclude that $g^{*}=0$. (Surjective) Let any $\left(\ell^{*}, t^{*}\right) \in$ $V^{*} \oplus W^{*}$, then define the function $f^{*} \in(V \oplus W)^{*}$ such that $f^{*}(v, w)=\ell^{*}(v)+t^{*}(w)$, then $\left.f^{*}\right|_{V \oplus\{0\}}=\ell^{*}$ and $\left.f^{*}\right|_{\{0\} \oplus W}=t^{*}$, therefore $\Phi\left(f^{*}\right)=\left(\ell^{*}, t^{*}\right)$.

Definition 5.7.6. Let $f: V \rightarrow W$ be a $k$-linear morphism. We define the dual (or transpose) $k$-linear morphism of $f$ as $f^{*}=f^{T}: W^{*} \rightarrow V^{*}$. This is a particular case of the induced Mor $k$-linear morphism, by taking the $k$-vector space $k$, so that

$$
f^{*}: \operatorname{Mor}(W, k) \rightarrow \operatorname{Mor}(V, k), \alpha \mapsto \alpha f .
$$

It is immediate from this definition that $(g f)^{*}=f^{*} g^{*}$ for every composable pair of functionals $f$ and $g$.

Proposition 5.7.7. Let $f: V \rightarrow W$ be a $k$-linear morphism and consider its dual $f^{*}: W^{*} \rightarrow$ $V^{*}$. If $f$ is injective, then $f^{*}$ is surjective, on the other hand, if $f$ is surjective, then $f^{*}$ is injective. Hence, if $f$ is an isomorphism, then $f^{*}$ is an isomorphism.

Proposition 5.7.8. Let $V_{1} \rightarrow V_{2} \rightarrow V_{3}$ be an exact sequence, then $V_{3}^{*} \rightarrow V_{2}^{*} \rightarrow V_{1}^{*}$ is also exact.

Proof. Let $f: V_{1} \rightarrow V_{2}$ and $g: V_{2} \rightarrow V_{3}$ be such that $\operatorname{im}(f)=\operatorname{ker}(g)$. Then, given any $\alpha \in V_{3}^{*}$ and consider $g^{*}(\alpha)=\alpha g$, we then have $f^{*}(\alpha g)=(\alpha g) f=\alpha(g f)=0$ from the construction of $f, g$, thus $\operatorname{im}\left(g^{*}\right) \subseteq \operatorname{ker}\left(f^{*}\right)$. On the other hand, let $\beta \in \operatorname{ker}\left(f^{*}\right) \subseteq V_{2}^{*}$ so that $f^{*}(\beta)=\beta f=0$, which is the same thing as requiring $\left.\beta\right|_{\operatorname{im}(f)}=0$; this way we can see that for any element $\alpha \in V_{3}^{*}$ we have $g^{*}(\alpha)=\alpha g$ be such that $\left.\alpha g\right|_{\operatorname{im}(f)}=0$ since $\operatorname{ker}(g)=\operatorname{im}(f)$ and thus already satisfies as an element of the kernel of $f^{*}$, therefore $\operatorname{ker}\left(f^{*}\right) \subseteq \operatorname{im}\left(g^{*}\right)$.

Proposition 5.7.9. Let $V_{1} \rightarrow V_{2} \rightarrow V_{3}$ be exact sequence of $k$-vector spaces, then for any given $k$-vector space $W$ we have that

$$
\operatorname{Mor}\left(V_{3}, W\right) \longrightarrow \operatorname{Mor}\left(V_{2}, W\right) \longrightarrow \operatorname{Mor}\left(V_{1}, W\right)
$$

is exact (contravariant Mor).
Proof. The proof is almost identical to the latter.
Proposition 5.7.10. Let $V_{1} \rightarrow V_{2} \rightarrow V_{3}$ be exact sequence of $k$-vector spaces, then for any given $k$-vector space $W$ we have that

$$
\operatorname{Mor}\left(W, V_{1}\right) \longrightarrow \operatorname{Mor}\left(W, V_{2}\right) \longrightarrow \operatorname{Mor}\left(V_{3}, W\right)
$$

is exact (covariant Mor).
Proof. The proof only diverges from the above because we have the mappings $g_{*}(\alpha)=$ $g \alpha$ and $f_{*}(\beta)=f \beta$.

Proposition 5.7.11. Consider a $k$-linear morphism $f: V \rightarrow W$ such that $\operatorname{rank}(f)$ is finite (that is, $\operatorname{dim}_{k}(\operatorname{im}(f))$ exists). Then $\operatorname{rank}(f)=\operatorname{rank}\left(f^{*}\right)$.

Proof. Notice that if $f: V \rightarrow U \mapsto W$ is the decomposition of $f$, then we can consider the dual $f^{*}: W^{*} \rightarrow V^{*}$ as having a decomposition $W^{*} \rightarrow U^{*} \mapsto V^{*}$, since the dual of an injective map is surjective and the dual of a surjective map is injective. Therefore, given a basis $B$ for $U$, since $\forall b \in B$ there exists $v \in V$ for which $f(v)=b$ we can say that $b \in \operatorname{im}(f)$ and, moreover, $\operatorname{rank}(f) \geqslant \operatorname{dim}_{k}(U)$. Also, since $\operatorname{im}(f) \subseteq U$, then we have directly that $\operatorname{rank}(f) \leqslant \operatorname{dim}_{k}(U)$, thus $\operatorname{rank}(f)=\operatorname{dim}_{k}(U)$. The same shows that $\operatorname{rank}\left(f^{*}\right)=\operatorname{dim}_{k}\left(U^{*}\right)$ and since $U \simeq U^{*}$ we find that $\operatorname{dim}_{k}(U)=\operatorname{dim}_{k}\left(U^{*}\right)$ and hence $\operatorname{rank}(f)=\operatorname{rank}\left(f^{*}\right)$.

Proposition 5.7.12. Let $\ell \in \operatorname{Mor}(V, W)$, where $V \simeq k^{n}$ and $W \simeq k^{m}$, so that $\ell$ can be written as a matrix $L: k^{n} \rightarrow k^{m}$. Then the dual $\ell^{*}$ is represented by the matrix $L^{T}$.

Proof. Define $\left(v_{j}\right)_{j=1}^{n}$ an ordered basis for $V$ and $\left(v_{i}^{*}\right)_{i=1}^{n}$ be its dual ordered basis; define also $\left(w_{i}\right)_{i=1}^{m}$ to be an ordered basis for $W$ and $\left(w_{j}^{*}\right)_{j=1}^{m}$ be its dual ordered basis. Notice that the dual of the matrix, that is, $L^{*}:\left(k^{m}\right)^{*} \rightarrow\left(k^{n}\right)^{*}$ can be written as $L^{*}: k^{m} \rightarrow k^{n}$, since $\left(k^{m}\right)^{*} \simeq k^{m}$ and $\left(k^{n}\right)^{*} \simeq k^{n}$. Define the representations

$$
L=\left[\begin{array}{ccc}
t_{1,1} & \ldots & t_{1, n} \\
\vdots & \ddots & \vdots \\
t_{m, 1} & \ldots & t_{m, n}
\end{array}\right] \text { and define } L^{*}=\left[\begin{array}{ccc}
d_{1,1} & \ldots & d_{1, m} \\
\vdots & \ddots & \vdots \\
d_{n, 1} & \ldots & d_{n, m}
\end{array}\right]
$$

By definition of the dual linear morphism, we have $L^{*}\left(w_{j}^{*}\right)=w_{j}^{*} L$. When this transformation is thought of as a $k^{n} \rightarrow k^{m}$ matrix, the definition of a matrix for a linear transformation with respect to the given dual basis for $W^{*}$ and $V^{*}$ is defined such that $L^{*}\left(w_{j}^{*}\right)=\sum_{r=1}^{n} d_{r, j} v_{r}^{*}$. Hence, given any element $v_{k} \in\left(v_{j}\right)_{j=1}^{n} \subseteq V$, from the first definition:

$$
\begin{equation*}
w_{j}^{*} L\left(v_{k}\right)=w_{j}^{*}\left(\sum_{r=1}^{m} t_{r, k} w_{r}\right)=\sum_{r=1}^{m} t_{r, k} w_{j}^{*}\left(w_{r}\right)=t_{j, k} \tag{5.2}
\end{equation*}
$$

from the fact that we defined $w_{j}^{*}$ so that $w_{j}^{*}\left(w_{j}\right)=1$ and $w_{j}^{*}\left(w_{i}\right)=0$ for $i \neq j$. Now, from the second definition that we discussed, we have (applying the same $v_{k}$ ):

$$
\begin{equation*}
\sum_{r=1}^{n} d_{r, j} v_{r}^{*}\left(v_{k}\right)=d_{k, j} \tag{5.3}
\end{equation*}
$$

from the same fact. Thus we conclude that $d_{k, j}=t_{j, k}$ for all $1 \leqslant k \leqslant n$ and $1 \leqslant j \leqslant m$, thus indeed $L^{*}=L^{T}$

Definition 5.7.13 (Column and Row Rank). Let $L: k^{n} \rightarrow k^{m}$ be a matrix. Define $\left(v_{j}\right)_{j=1}^{n}$ to be the vectors whose components are given by the ordered $j$-th column elements of the matrix representation of $L$. Analogously, define $\left(w_{i}\right)_{i=1}^{m}$ to be the vectors whose components are the ordered $i$-th row elements of the matrix representation of $L$. We define the column rank of $L$ to be $\operatorname{dim}_{k}\left(\operatorname{span}\left(v_{j}\right)_{j=1}^{n}\right)$ and the row rank of $L$ to be $\operatorname{dim}_{k}\left(\operatorname{span}\left(w_{i}\right)_{i=1}^{m}\right)$.

Proposition 5.7.14. Let $L: V \rightarrow W$ be a linear morphism with $V \simeq k^{n}$ and $W \simeq k^{m}$, its row and column rank of its matrix representation $k^{n} \rightarrow k^{m}$ are both equal to the rank of $L$.

Proof. Let $\left(v_{j}\right)_{j=1}^{n}$ and $\left(w_{i}\right)_{i=1}^{m}$ be the column and row vectors of the matrix representation of L. (Rank equals column rank) Let $B_{V}$ and $S_{W}$ be ordered basis for $V$ and $W$ respectively and any element $w \in \operatorname{im}(L)=\operatorname{span}\left(L b_{k}: b_{k} \in B_{V}\right)$ notice that since $L b_{k}=$ $\sum_{i=1}^{m} t_{i, k} s_{i}$, where $t_{i, j}$ are the elements of the matrix representation of $L$, then we see that $L b_{k}$ is a linear combination of the $k$-th column vector of $L$, hence $L b_{k} \in \operatorname{span}\left(v_{j}\right)_{j=1}^{n}$ for all $1 \leqslant k \leqslant m$. Therefore we conclude that $w \in \operatorname{span}\left(v_{i}\right)_{i=1}^{n}$. Moreover, we have $\operatorname{span}\left(v_{i}\right)_{i=1}^{n} \subseteq \operatorname{im}(L)$, thus $\operatorname{span}\left(v_{i}\right)_{i=1}^{n}=\operatorname{im}(L)$, hence we conclude finally that $\operatorname{dim}_{k}(\operatorname{im}(L))=\operatorname{rank}(L)=\operatorname{dim}_{k}\left(\operatorname{span}\left(v_{i}\right)_{i=1}^{n}\right)$ which states that the rank of $L$ equals the
column rank. (Rank equals column rank) Notice that $L^{*}:\left(k^{m}\right)^{*} \rightarrow\left(k^{n}\right)^{*}$ is isomorphic to the matrix $k^{m} \rightarrow k^{n}$ and the matrix representation of the dual of $L$ is equal to its transpose by 5.7.12, moreover we proved lastly that the column rank of a matrix equals its rank, thus it should be true that $\operatorname{rank}\left(L^{*}\right)=\operatorname{dim}_{k}\left(\operatorname{span}\left(w_{i}\right)_{i=1}^{m}\right)$. Since $L^{*} \simeq L$, it follows that $\operatorname{rank}(L)=\operatorname{dim}_{k}\left(\operatorname{span}\left(w_{i}\right)_{i=1}^{m}\right)$. Hence we conclude that

$$
\operatorname{rank}(L)=\operatorname{rank}(\text { column })=\operatorname{rank}(\text { row }) .
$$

Definition 5.7.15 (Bilinear map). A map $V^{*} \times V \rightarrow k$ that has the mapping $(\alpha, v) \mapsto \alpha(v)$ is called a bilinear map. This map is linear only when we fix the second component to some $v \in V$ so that $f_{v}: V^{*} \rightarrow k$ with the mapping $\alpha \mapsto \alpha(v)$ is a $k$-linear morphism.

This defines a rather interesting canonical $k$-linear morphism

$$
\Psi_{V}: V \rightarrow\left(V^{*}\right)^{*}, \text { mapping } v \mapsto\left(f_{v}=(\alpha \mapsto \alpha(v))\right)
$$

Proposition 5.7.16. The morphism $\Psi_{V}$ is an injection for every given vector space $V$.
Proposition 5.7.17. A finite dimensional space $V$ is isomorphic to its double dual $V^{* *}$. The same can't be said if $V$ is infinite dimensional.

Proof. Let $\left\{v_{i}\right\}_{i=1}^{n}$ be a basis for $V,\left\{v_{i}^{*}\right\}_{i=1}^{n}$ be a basis for $V^{*}$ dual to the first given basis, and $\left.\left\{v_{i}^{* *}\right)_{i=1}^{n}\right\}$ be a basis for $V^{* *}$ dual to the second given basis. We'll show that $\Psi_{V}\left(v_{i}\right)=v_{i}^{* *}$. Let $v_{i}$ be an element of the given basis of $V$, then, we know from the definition of the dual of the basis that $v_{k}^{*}\left(v_{i}\right)=\delta_{i, k}$ and $v_{t}^{* *}\left(v_{k}^{*}\right)=\delta_{k, t}$. Moreover, since $\Psi_{V}$ is an injection, we see that if $\Psi_{V}\left(v_{i}\right) \neq 0$, hence $v_{i} \mapsto v_{i}^{* *}$, which proves the statement.

Definition 5.7.18 (Annihilator). Let $V$ be a finite dimensional $k$-vector space and $W$ be a subspace of $V$. The annihilator of $W$ is defined as the subspace of $V^{*}$ given by

$$
W^{0}=\left\{\alpha \in V^{*}: \alpha(W)=0\right\}
$$

Proposition 5.7.19. Let $V$ be a finite dimensional space and $W \subseteq V$ a subspace. Then

$$
\operatorname{dim}_{k}(V)=\operatorname{dim}_{k}(W)+\operatorname{dim}_{k}\left(W^{0}\right)
$$

Proof. Let the inclusion $t: W \hookrightarrow V$, then its dual $\iota^{*}: V^{*} \rightarrow W^{*}$ is surjective. Moreover, let $\alpha \in W^{0}$ then, $\iota^{*}(\alpha)=\alpha \iota$, since $\operatorname{im}(\iota) \subseteq W$ we conclude that $\iota^{*}(\alpha)=0$ and thus $\alpha \in \operatorname{ker}\left(\iota^{*}\right)$, so that $W^{0} \subseteq \operatorname{ker}\left(\iota^{*}\right)$. Moreover, if $\alpha \in \operatorname{ker}\left(\iota^{*}\right)$ then, in particular, $\alpha(W)=0$ hence $\alpha \in W^{0}$. Then we find that $W^{0}=\operatorname{ker}\left(\iota^{*}\right)$. We now use the theorem 5.6.23 on the morphism $\iota^{*}$ so that $\operatorname{dim}_{k}\left(V^{*}\right)=\operatorname{dim}_{k}\left(\operatorname{ker}\left(\iota^{*}\right)\right)+\operatorname{rank}\left(\iota^{*}\right)$. Since $\iota^{*}$ is surjective, then $\operatorname{im}\left(\iota^{*}\right)=W^{*}$ and therefore $\operatorname{rank}\left(\iota^{*}\right)=\operatorname{dim}_{k}\left(W^{*}\right) ;$ moreover, $\operatorname{dim}_{k}\left(\operatorname{ker}\left(\iota^{*}\right)\right)=\operatorname{dim}_{k}\left(W^{0}\right)$; and finally $\operatorname{dim}_{k}\left(V^{*}\right)=\operatorname{dim}_{k}(V)$, and $\operatorname{dim}_{k}\left(W^{*}\right)=\operatorname{dim}_{k}(W)$. Thus indeed $\operatorname{dim}_{k}(V)=$ $\operatorname{dim}_{k}(W)+\operatorname{dim}_{k}\left(W^{0}\right)$.

Complexification and decomplex

## Chapter 6

## Multilinear Algebra

## Add content about bilinear maps

### 6.1 Multilinear Maps and Tensor Products

The goal of this section is to introduce another concept of linearity, one of which is induced by a kind of map that is called multilinear. This is a very important generalization of what we've been studying so far.

Definition 6.1.1 (Multilinear map). Let $\left\{V_{i}\right\}_{i=1}^{n}$ be a collection of $k$-vector spaces and $W$ be a $k$-vector space. We say that a map

$$
\varphi: \prod_{i=1}^{n} V_{i} \rightarrow W
$$

is a multilinear map if for a given $1 \leqslant t \leqslant n$ we have that, for all collections $\left\{v_{i}: v_{i} \in\right.$ $\left.V_{i}, i \neq t\right\}$, the map

$$
\varphi\left(v_{1}, \ldots, v_{t-1},-, v_{t+1}, \ldots, v_{n}\right): V_{t} \longrightarrow W, \text { mapping } x \stackrel{\varphi}{\longmapsto} \varphi\left(v_{1}, \ldots, v_{t-1}, x, v_{t}, \ldots, v_{n}\right)
$$

is a $k$-linear morphism.
Multilinear structures seems interesting and what is even better is that there is a way of working with multilinear maps by means of the machinery acquired when we where studying linear structures - we can simply linearize a multilinear map! That is a great idea, one that is accomplished via the introduction of a structure called tensor product. We'll now focus on the construction of it.

Given a finite collection of $k$-vector spaces $\left\{V_{i}\right\}_{i \in I}$, we define the free $k$-vector space $\mathcal{M}=k^{\oplus} \prod_{i \in I} V_{i}$ - that is, the space of set-functions $\prod_{i \in I} V_{i} \rightarrow k$ with finite support. A map $f \in \mathcal{M}$ is entirely described by the set of all tuples of its domain. For each tuple $\left(v_{i}\right)_{i \in I} \in \prod_{i \in I} V_{i}$, we define a characteristic map $\delta_{\left(v_{i}\right)_{i \in I}}: \prod_{i \in I} V_{i} \rightarrow\{0,1\}-$ which happens to be an element of the space $\mathcal{M}$. Since

$$
f(v)=\sum_{u \in \prod_{i \in I} V_{i}} f(u) \delta_{v}(u)=f(v) \delta_{v}(v)=f(v)
$$

we conclude that the collection $\left\{\delta_{v} \in \mathcal{M}: v \in V\right\}$ is a basis for the space $\mathcal{M}$.
For the sake of brevity, we'll simply denote $\delta_{v}=v$ for every $v \in \prod_{i \in I} V_{i}$. The space $\mathcal{M}$ can then be viewed as the space of formal linear combination of the tuples of $\prod_{i \in I} V_{i}$, that is

$$
\mathcal{M}=\left\{\sum a_{v} v: v \in \prod_{i \in I} V_{i} \text { and } a_{v} \in k\right\} .
$$

Thus, if char $k=0$ and exists $i \in I$ such that $\operatorname{dim}_{k}\left(V_{i}\right)>0$, then $\mathcal{M}$ is an infinitedimensional vector space.

The next step in our construction is to refine the space $\mathcal{M}$ by means of taking the quotient of $\mathcal{M}$ by some subspace $\mathcal{M}_{0}$ so that the maps $f \in \mathcal{M} / \mathcal{M}_{0}$ satisfy the properties arising from the multilinear structure induced by the maps Definition 6.1.1. We can do so by considering the collection of maps of the following form

$$
f_{0}=\left(v_{1}, \ldots, v_{j}+a v_{j}^{\prime}, \ldots, v_{n}\right)-\left(v_{1}, \ldots, v_{j}, \ldots, v_{n}\right)-a\left(v_{1}, \ldots, v_{j}^{\prime}, \ldots, v_{n}\right) \in \mathcal{M} .
$$

We may simply define $\mathcal{M}_{0} \subseteq \mathcal{M}$ to be the subspace spanned by the maps of the same form as $f_{0} \in \mathcal{M}$. Now, when we take the quotient $\mathcal{M} / \mathcal{M}_{0}$ we essentially classify the maps of the form $f_{0}$ as the zero-maps - so that every element of $\mathcal{M} / \mathcal{M}_{0}$ satisfy an internal version of the properties of the multilinear maps. We are now ready to define what we call a tensor product - we actually already constructed it, is the space $\mathcal{M} / \mathcal{M}_{0}$ !

Definition 6.1.2 (Tensor product). We define the tensor product of the finite collection of $k$-vector spaces $\left\{V_{i}\right\}_{i \in I}$ as the quotient space

$$
\bigotimes_{i \in I} V_{i}=\mathcal{M} / \mathcal{M}_{0}
$$

where the elements of $\bigotimes_{i \in I} V_{i}$ are called tensors, and in special, the elements of the form

$$
\otimes_{i \in I} v_{i}=\left(v_{i}\right)_{i \in I}+\mathcal{M}_{0} \in \bigotimes_{i \in I} V_{i}
$$

are called factorizable tensors.
Lemma 6.1.3. Let $\left\{V_{i}\right\}_{i \in I}$ be a finite collection of $k$-vector spaces. The canonical map

$$
\otimes: \prod_{i \in I} V_{i} \rightarrow \bigotimes_{i \in I} V_{i}, \text { mapping }\left(v_{i}\right)_{i \in I} \longmapsto \otimes_{i \in I} v_{i}
$$

is multilinear.
Proof. Let $|I|=n$, and $a \in k$ be any constant, then

$$
\begin{aligned}
\otimes\left(v_{1}, \ldots, v_{j}+a v_{j}^{\prime}, \ldots, v_{n}\right) & =v_{1} \otimes \cdots \otimes\left(v_{j}+a v_{j}^{\prime}\right) \otimes \cdots \otimes v_{n} \\
& =\left(v_{1}, \ldots, v_{j}+a v_{j}^{\prime}, \ldots, v_{n}\right)+\mathcal{M}_{0} \\
& =\left[\left(v_{1}, \ldots, v_{j}, \ldots, v_{n}\right)+\mathcal{M}_{0}\right]+\left[a\left(v_{1}, \ldots, v_{j}^{\prime}, \ldots, v_{n}\right)+\mathcal{M}_{0}\right] \\
& =v_{1} \otimes \cdots \otimes v_{j} \otimes \cdots \otimes v_{n}+a\left(v_{1} \otimes \cdots \otimes v_{j}^{\prime} \otimes \cdots \otimes v_{n}\right) \\
& =\otimes\left(v_{1}, \ldots, v_{j}, \ldots, v_{n}\right)+a \otimes\left(v_{1}, \ldots, v_{j}^{\prime}, \ldots, v_{n}\right) .
\end{aligned}
$$

Where we used the fact that

$$
\left(v_{1}, \ldots, v_{j}+a v_{j}^{\prime}, \ldots, v_{n}\right)=\left(v_{1}, \ldots, v_{j}, \ldots, v_{n}\right)+a\left(v_{1}, \ldots, v_{j}^{\prime}, \ldots, v_{n}\right) \in \mathcal{M} / \mathcal{M}_{0}
$$

Theorem 6.1.4 (Universal property of tensor products). Let $\left\{V_{i}\right\}_{i \in I}$ be a finite collection of $k$-vector spaces, and $L$ be any $k$-vector space, and $f: \prod_{i \in I} V_{i} \rightarrow L$ be any multilinear map. Then there exists a unique $k$-linear morphism $\ell: \otimes_{i \in I} V_{i} \rightarrow L$ such that the following diagram commutes


Proof. Let $|I|=n$. (Uniqueness) Suppose that for any $L$ and $f$, the morphism $\ell$ exists, so that $f=\ell \circ \otimes$ and therefore

$$
\ell \otimes\left(v_{1}, \ldots, v_{n}\right)=\ell\left(v_{1} \otimes \cdots \otimes v_{n}\right)=f\left(v_{1}, \ldots, v_{n}\right)
$$

then certainly $\ell$ is uniquely defined by the pointwise image of $f$, since $\left\{\left(v_{1}, \ldots, v_{n}\right) \in\right.$ $\left.\prod_{i=1}^{n} V_{i}\right\}$ generates $\bigotimes_{i=1}^{n} V_{i}$.
(Existence) Let $g: \mathcal{M} \rightarrow L$ be defined as $g\left(v_{1}, \ldots, v_{n}\right)=f\left(v_{1}, \ldots, v_{n}\right)$ so that $f$ completely determines the image of $g$. Let now $\left(v_{1}, \ldots, v_{j-1}, 0, v_{j+1}, \ldots, v_{n}\right) \in \mathcal{M}_{0}$, then

$$
g\left(v_{1}, \ldots, v_{j-1}, 0, v_{j+1}, \ldots, v_{n}\right)=f\left(v_{1}, \ldots, v_{j-1}, 0, v_{j+1}, \ldots, v_{n}\right)=0
$$

because $f$ is a multilinear map, then $\left(v_{1}, \ldots, v_{j-1}, 0, v_{j+1}, \ldots, v_{n}\right) \in \operatorname{ker}(g)$. Moreover, for the second type of element of $\mathcal{M}_{0}$

$$
\begin{aligned}
g\left(a\left(v_{1}, \ldots, v_{j-1}, 0, v_{j+1}, \ldots, v_{n}\right)\right) & =f\left(a\left(v_{1}, \ldots, v_{j-1}, 0, v_{j+1}, \ldots, v_{n}\right)\right) \\
& =f\left(a\left(v_{1}, \ldots, v_{j-1}, 0, v_{j+1}, \ldots, v_{n}\right)\right) \\
& =a f\left(v_{1}, \ldots, v_{j-1}, 0, v_{j+1}, \ldots, v_{n}\right) \\
& =0
\end{aligned}
$$

thus $a\left(v_{1}, \ldots, v_{j-1}, 0, v_{j+1}, \ldots, v_{n}\right) \in \operatorname{ker}(g)$ and hence $\mathcal{M}_{0} \subseteq \operatorname{ker}(g)$. From the universal property of quotients Theorem 5.4.3 we have that $g$ induces the uniquely defined $k$-linear morphism $\ell: \mathcal{M} / \mathcal{M}_{0} \rightarrow L$ such that the diagram commutes

then $g=\ell \pi$ and therefore for all $\left(v_{1}, \ldots, v_{n}\right) \in \prod_{i=1}^{n} V_{i}$ we have

$$
f\left(v_{1}, \ldots, v_{n}\right)=g\left(v_{1}, \ldots, v_{n}\right)=\ell \pi\left(v_{1}, \ldots, v_{n}\right)=\ell\left(v_{1} \otimes \cdots \otimes v_{n}\right)
$$

then indeed $\ell \otimes=f$ as wanted.

Corollary 6.1.5 (Multilinear maps are isomorphic to linear maps). Let $\left\{V_{i}\right\}_{i=1}^{p}$ be a finite collection of $k$-vector spaces, and $\operatorname{Mor}\left(V_{1}, \ldots, V_{p} ; L\right)$ denote the space of multilinear $\operatorname{maps} \prod_{i=1}^{p} V_{i} \rightarrow L$, where $L$ is a given $k$-vector space. The $k$-linear morphism

$$
\psi: \operatorname{Mor}\left(V_{1}, \ldots, V_{p} ; L\right) \longrightarrow \operatorname{Mor}\left(\bigotimes_{i=1}^{p} V_{i}, L\right), f \stackrel{\psi}{\longmapsto} \ell
$$

where $f=\ell \otimes$ (as in Theorem 6.1.4), is an isomorphism, so that

$$
\operatorname{Mor}\left(V_{1}, \ldots, V_{p} ; L\right) \simeq \operatorname{Mor}\left(\bigotimes_{i=1}^{p} V_{i}, L\right)
$$

Proof. Let $\ell \in \operatorname{Mor}\left(\bigotimes_{i=1}^{p} V_{i}, L\right)$, then $f=\ell \otimes \in \psi^{-1}(\ell)$ thus $\psi$ is surjective. On the other hand, if $f \neq 0$, then $\ell \otimes \neq 0$ and in particular $\ell \neq 0$ hence $\psi(f)=\ell \neq 0$, therefore $\operatorname{ker}(\psi)=0$. Since $\psi$ is linear, then $\psi$ is injective.

## Dimensions and Bases of Tensor Products

Proposition 6.1.6. Let $\left\{V_{i}\right\}_{i \in I}$ be a finite collection of $k$-vector spaces. If exists $j \in I$ such that $V_{j}=0$ is the null space then $\bigotimes_{i \in I} V_{i}=0$, that is, the tensor product is the null space.
Proof. Let $V_{j}=0$ for some $j \in I$. Consider any $f \in \operatorname{Mor}\left(\prod_{i \in I} V_{i}, L\right)$, where $L$ is a $k$-vector space. Since $f$ is linear on $V_{j}$, consider any element $\left(v_{i}\right)_{i \in I} \in \prod_{i \in I} V_{i}$ then necessarily $v_{j}=0$ and hence $f\left(\left(v_{i}\right)_{i \in I}\right)=0$, since $V_{j}$ has only one element, every element of the domain of $f$ is mapped to zero, which implies in $f=0$. Consider in particular the mapping $\otimes \in \operatorname{Mor}\left(\prod_{i \in I} V_{i}, \bigotimes_{i \in I} V_{i}\right)$ (Lemma 6.1.3) then from the latter discussion we have $\otimes=0$, since $\bigotimes_{i \in I} V_{i}=\operatorname{span}(\operatorname{im}(\otimes))$, then we conclude that $\bigotimes_{i \in I} V_{i}=0$.
Proposition 6.1.7 (Tensor product dimension). The dimension of the tensor product of a finite collection of finite $k$-vector spaces $\left\{V_{i}\right\}_{i \in I}$ is equal to the product of the dimensions of the vector spaces, that is

$$
\operatorname{dim}_{k} \bigoplus_{i \in I} V_{i}=\prod_{i \in I} \operatorname{dim}_{k} V_{i}
$$

Proof. Let $|I|=n$. If exists a null vector space in the collection, then the dimension of the tensor product is null, thus the proposition follows. Suppose there is no such null vector space in the collection, then since $\bigotimes_{i \in I} V_{i} \simeq\left(\otimes_{i \in I} V_{i}\right)^{*}=\operatorname{Mor}\left(\bigotimes_{i \in I} V_{i}, k\right)$. From Corollary 6.1.5 we find that $\left(\bigotimes_{i \in I} V_{i}\right)^{*} \simeq\left(\prod_{i \in I} V_{i}\right)^{*}$. Let $B_{i}=\left\{e_{j}^{(i)}\right\}_{j=1}^{\operatorname{dim}_{k} V_{i}}$ be a basis for $V_{i}$ for every $i \in I$, so that for every element $v_{i} \in V_{i}$ we can write it as a linear combination $v_{i}=\sum_{j=1}^{\left|B_{1}\right|} x_{j}^{(i)} e_{j}^{(i)}$ where $x_{j}^{(i)} \in k$. Then, for any multilinear map $f \in\left(\prod_{i \in I} V_{i}\right)^{*}$ we have

$$
f\left(v_{1}, \ldots, v_{n}\right)=f\left(\sum_{j_{1}=1}^{\left|B_{1}\right|} x_{j_{1}}^{(1)} e_{j_{1}}^{(1)}, \ldots, \sum_{j_{n}=1}^{\left|B_{n}\right|} x_{j_{n}}^{(n)} e_{j_{n}}^{(n)}\right)=\sum_{\substack{1 \leqslant j_{i} \leqslant\left|B_{i}\right|, 1 \leqslant i \leqslant n}} x_{j_{1}}^{(1)} \ldots x_{j_{n}}^{(n)} f\left(e_{j_{1}}^{(1)}, \ldots, e_{j_{n}}^{(n)}\right)
$$

so that the collection $\prod_{i \in I} B_{i}$ forms a base for $\left(\prod_{i \in I} V_{i}\right)^{*}$ and therefore

$$
\operatorname{dim}_{k}\left(\prod_{i \in I} V_{i}\right)^{*}=\prod_{i \in I}\left|B_{i}\right|=\prod_{i \in I} \operatorname{dim}_{k} V_{i}
$$

Hence, since $\bigoplus_{i \in I} V_{i} \simeq\left(\prod_{i \in I} V_{i}\right)$, we conclude that $\operatorname{dim}_{k} \bigoplus_{i \in I} V_{i}=\prod_{i \in I} \operatorname{dim}_{k} V_{i} . \quad$ घ
Lemma 6.1.8 (Tensor basis). Given a finite collection of $k$-vector spaces $\left\{V_{i}\right\}_{i \in I}$ and suppose none of them are null. The basis for the tensor product $\bigotimes_{i \in I} V_{i}$ is given by the collection of factorizable tensors $\left\{\otimes_{i \in I} e_{j_{i}}^{(i)}\right\}$, where $e_{j_{i}}^{(i)} \in B_{i}$ and $B_{i}$ is a basis for $V_{i}$.

What about infinite dimensional vector spaces and infinite tensor products?

## Tensor Product of Function Spaces

Proposition 6.1.9. Let $\left\{S_{i}\right\}_{i \in I}$ be a finite collection of sets. There exists a canonical isomorphism

$$
k^{\oplus\left(\prod_{i \in I} S_{i}\right)} \simeq \bigotimes_{i \in I} k^{\oplus S_{i}}
$$

Proof. We show that the $k$-linear morphism $\phi: k^{\oplus\left(\prod_{i \in I} S_{i}\right)} \rightarrow \bigoplus_{i \in I} k^{\oplus S_{i}}$ mapping $\delta_{\left(s_{i}\right)_{i \in I}} \mapsto$ $\otimes_{i \in I} \delta_{s_{i}}$, where $\delta$ is the Kronecker delta function, is an isomorphism. First, from Proposition 5.3 .5 we find that $\left\{\delta_{\left(s_{i}\right)_{i \in I}}:\left(s_{i}\right)_{i \in I} \in \prod_{i \in I} S_{i}\right\}$ is a base for the space $k^{\oplus\left(\prod_{i \in I} S_{i}\right)}$ and, in particular, $\left\{\delta_{s_{i}}: s_{i} \in S_{i}\right\}$ is a base for the space $k^{\oplus S_{i}}$. From Lemma 6.1 .8 we find that $\left\{\otimes_{i \in I} \delta_{s_{i}}: s_{i} \in S_{i}\right\}$ is a basis for $\bigotimes_{i \in I} k^{\oplus S_{i}}$.

### 6.2 Canonical Isomorphisms and Tensor Products

## Associativity and Commutativity

Proposition 6.2.1 (Associativity). Let $k$-vector spaces $V, W, L$, the map $(V \otimes W) \otimes L \rightarrow V \otimes(W \otimes L)$ mapping $(v \otimes w) \otimes \ell \mapsto v \otimes(w \otimes \ell)$
is a canonical isomorphism. Hence for any collection of $k$-vector spaces $\left\{V_{i}\right\}_{i=1}^{p}$ we can write any arrangement of parenthesis for their tensor product as canonically isomorphic to $V_{1} \otimes \cdots \otimes V_{p}$.

Proof. Let $B_{V}, B_{W}, B_{L}$ be basis for the vector spaces $V, W, L$ respectively. Notice that $\left\{(v \otimes w) \otimes \ell:(v, w, \ell) \in B_{V} \times B_{W} \times L\right\}$ is a basis for $(V \otimes W) \otimes L$ and $\{v \otimes(w \otimes \ell)$ : $\left.(v, w, \ell) \in B_{V} \times B_{W} \times B_{L}\right\}$ is a basis for $V \otimes(W \otimes L)$ (from Lemma 6.1.8). Therefore, the $\operatorname{map}(v \otimes w) \otimes \ell \mapsto v \otimes(w \otimes \ell)$ transforms one base into another, which implies that it is a canonical isomorphism of the considered spaces.

Proposition 6.2.2 (Commutativity). Let $\left\{V_{i}\right\}_{i=1}^{p}$ be a collection of $k$-vector spaces, and $\sigma$ be any permutation of the numbers $\{1, \ldots, p\}$. Define the $k$-linear morphism

$$
f_{\sigma}: \bigotimes_{i=1}^{p} V_{i} \rightarrow \bigotimes_{i=1}^{p} V_{\sigma(i)} \text { mapping } v_{1} \otimes \cdots \otimes v_{p} \mapsto v_{\sigma(1)} \otimes \cdots \otimes v_{\sigma(p)}
$$

where $f_{\tau \sigma}=f_{\tau} f_{\sigma}$, for any permutation $\tau$ on $\{1, \ldots, p\}$. Then $f_{\sigma}$ is a canonical isomorphism.

Proof. Let the map $g_{\sigma}: \prod_{i=1}^{p} V_{i} \rightarrow \bigotimes_{i=1}^{p} V_{\sigma(i)}$ such that $\left(v_{1}, \ldots, v_{p}\right) \mapsto v_{\sigma(1)} \otimes \cdots \otimes v_{\sigma(p)}$. Then the following diagram commutes


Now, by means of Theorem 6.1.4, we find that the morphism $f_{\sigma}$ is unique. Notice that $f_{\sigma}$ maps the base of $\bigotimes_{i=1}^{p} V_{i}$ to the base of $\bigotimes_{i=1}^{p} V_{\sigma(i)}$, hence $f_{\sigma}$ is an isomorphism.

## Duality

Proposition 6.2.3. Let $\left\{V_{i}\right\}_{i=1}^{p}$ be a collection of finite dimensional $k$-vector spaces. Then the $k$-linear morphism

$$
\bigoplus_{i=1}^{p} V_{i}^{*} \xrightarrow{\sim}\left[\bigoplus_{i=1}^{p} V_{i}\right]^{*} \text { mapping } f_{1} \otimes \cdots \otimes f_{p} \longmapsto\left(v_{1} \otimes \cdots \otimes v_{p} \mapsto f_{1}\left(v_{1}\right) \ldots f_{p}\left(v_{p}\right)\right)
$$

is a natural isomorphism.
Proof. Since $V_{i}$ is finite dimensional for all $i$, then $V_{i} \simeq V_{i}^{*}$, which in particular yield $\operatorname{dim}_{k}\left(V_{i}\right)=\operatorname{dim}_{k}\left(V_{i}^{*}\right)$ (Proposition 5.7.3), then $\operatorname{dim}_{k}\left(\bigoplus_{i} V_{i}^{*}\right)=\operatorname{dim}_{k}\left(\bigoplus_{i} V_{i}\right)$. Now, since $\bigotimes_{i} V_{i}$ is finite, then $\bigotimes_{i} V_{i} \simeq\left(\bigotimes_{i} V_{i}\right)^{*}$, which implies in $\operatorname{dim}_{k}\left(\bigotimes_{i} V_{i}\right)=$ $\operatorname{dim}_{k}\left(\bigotimes_{i} V_{i}\right)^{*}$. From this, $\operatorname{dim}_{k}\left(\bigoplus V_{i}^{*}\right)=\operatorname{dim}_{k}\left(\bigoplus_{i} V_{i}\right)^{*}$. From Corollary 5.6.25 it suffices to show that the map is surjective or injective. Let $f_{1} \otimes \cdots \otimes f_{p} \neq 0$, then its image maps to $f_{1}\left(v_{1}\right) \ldots f_{p}\left(v_{p}\right)$, which cannot be zero if $v_{1} \otimes \cdots \otimes v_{p} \neq 0$. This implies that the kernel of the $k$-linear map shown is zero, hence it's an isomorphism.

Proposition 6.2.4. Let $V, L$ be finite dimensional $k$-vector spaces. Then the $k$-linear morphism

$$
V^{*} \otimes L \xrightarrow{\sim} \operatorname{Mor}(V, L) \text { mapping } f \otimes \ell \longmapsto(v \mapsto f(v) \ell)
$$

is a canonical isomorphism.
Proof. Name the above morphism $\phi$. Let $\operatorname{dim}_{k}(V)=n$ and $\operatorname{dim}_{k}(L)=m$. Let basis $\left\{v_{j}\right\}_{j=1}^{n}$ and $\left\{\ell_{i}\right\}_{i=1}^{m}$ of $V$ and $L$, respectively. Then we find the corresponding dual basis
$\left\{v_{j}^{*}\right\}_{j=1}^{n}$ and $\left\{\ell_{i}^{*}\right\}_{i=1}^{m}$. Notice that $\phi: v_{j}^{*} \otimes \ell_{i} \mapsto\left(v \mapsto v_{j}^{*}(v) \ell_{i}\right)$, hence, given $g \in \operatorname{im}(\phi) \subseteq$ $\operatorname{Mor}(V, L)$ we can write its matrix representation $k^{n} \rightarrow k^{m}$ with factors $a_{i j}$ defined by

$$
g\left(v_{k}\right)=\sum_{i=1}^{m} a_{i k} \ell_{i}=v_{j}^{*}\left(v_{k}\right) \ell_{i}= \begin{cases}\ell_{i}, & k=j \\ 0, & \text { otherwise }\end{cases}
$$

for some $\left(v_{j}^{*}, \ell_{i}\right) \in\left\{v_{j}^{*}\right\}_{j=1}^{m} \times\left\{\ell_{i}\right\}_{i=1}^{m}$, so that $a_{i k}=0$ for all $k \neq j$ and $a_{i j}=1$. Notice that this makes $\operatorname{im}(\phi)$ a basis for $\operatorname{Mor}(V, L)$, transforming a basis into other, which qualifies $\phi$ as an isomorphism.

Notice that if $V$ is finite dimensional $k$-vector space, we can consider the special case of the endomorphism $\operatorname{End}(V) \simeq V^{*} \otimes V$. Notice that $\mathrm{id}_{V} \in \operatorname{End}(V)$ is such that, given $v_{j} \in\left\{v_{i}\right\}_{i=1}^{n}$, where the last is a basis for $V$, we have $\operatorname{id}_{V}\left(v_{j}\right)=\sum_{i=1}^{n} \delta_{i j} v_{i}=\sum_{i=1}^{n} v_{i}^{*}\left(v_{j}\right) v_{i}$. This implies in the mapping $\operatorname{id}_{V} \mapsto \sum_{i=1}^{n} v_{i}^{*} \otimes v_{i}$ for the canonical isomorphism.

Definition 6.2.5 (Trace). Given a finite $k$-vector space $V$, we define the canonical linear functional

$$
\operatorname{tr}: V^{*} \otimes V \rightarrow k, \alpha \otimes v \mapsto \alpha(v)
$$

Since $V^{*} \otimes V \simeq \operatorname{End}(V)$, then we can view the above definition in terms of endomorphisms, so that $\operatorname{tr}: \operatorname{End}(V) \rightarrow k$. For a more computational interpretation, let $\operatorname{dim}_{k} V=n$ and $\left\{v_{i}\right\}_{i=1}^{n}$ be a basis for $V$. Given a linear endomorphism $f \in \operatorname{End}(V)$, suppose that it's coefficients in the matrix representation are $a_{i j}$, so that for any $v_{k} \in\left\{v_{i}\right\}_{i=1}^{n}$ we have $f\left(v_{k}\right)=\sum_{i=1}^{n} a_{i j} v_{i}=\sum_{i, j=1}^{n} a_{i j} v_{j}^{*}\left(v_{k}\right) v_{i}$. From the canonical map described in Proposition 6.2.4 we find that $f \mapsto \sum_{i, j=1}^{n} a_{i j} v_{j}^{*} \otimes v_{i}$. Now, from the definition of the trace it follows that

$$
f \longmapsto \sum_{i, j=1}^{n} a_{i j} v_{j}^{*} \otimes v_{i} \stackrel{\operatorname{tr}}{\longmapsto} \sum_{i=1}^{n} a_{i i} .
$$

As one can note, this is the sum of the diagonal of the matrix representation of $f$ (which happens to be independent of the base, as can be asserted from the more abstract definition).

Corollary 6.2.6. Given finite dimensional vector spaces $V, L, W$ we have that

$$
\operatorname{Mor}(V \otimes L, W) \simeq \operatorname{Mor}(V, \operatorname{Mor}(L, W))
$$

Proof. Notice that from Proposition 6.2.4 we find

$$
\begin{aligned}
\operatorname{Mor}(V \otimes L, W) \simeq(V \otimes L)^{*} \otimes W & \simeq\left(V^{*} \otimes L^{*}\right) \otimes W \\
& \simeq V^{*} \otimes\left(L^{*} \otimes W\right) \\
& \simeq V^{*} \otimes \operatorname{Mor}(L, W) \\
& \simeq \operatorname{Mor}(V, \operatorname{Mor}(L, W))
\end{aligned}
$$

Corollary 6.2.7 (Tensor product of morphisms). Let $\left\{V_{i}\right\}_{i \in I}$ and $\left\{L_{i}\right\}_{i \in I}$ be a finite collection of finite $k$-vector spaces. Then the $k$-linear morphism

$$
\bigotimes_{i \in I} \operatorname{Mor}\left(V_{i}, L_{i}\right) \rightarrow \operatorname{Mor}\left(\bigotimes_{i \in I} V_{i}, \bigotimes_{i \in I} L_{i}\right), \otimes_{i \in I} f_{i} \mapsto\left(\otimes_{i \in I} v_{i} \mapsto \otimes_{i \in I} f_{i}\left(v_{i}\right)\right)
$$

is a canonical isomorphism.
Proof. First notice that

$$
\begin{aligned}
\operatorname{Mor}\left(\bigotimes_{i} V_{i}, \bigotimes_{i} L_{i}\right) \simeq\left(\bigotimes_{i} V_{i}\right)^{*} \otimes\left(\bigotimes_{i} L_{i}\right) & \simeq\left(\bigotimes_{i} V_{i}^{*}\right) \otimes\left(\bigotimes_{i} L_{i}\right) \\
& \simeq \bigotimes_{i} V_{i}^{*} \otimes L_{i} \\
& \simeq \bigotimes_{i} \operatorname{Mor}\left(V_{i}, L_{i}\right)
\end{aligned}
$$

from propositions 6.2.4, and 6.2.3, and 6.2.2. Moreover, the map is clearly injective and surjective, hence an isomorphism.

## Contractions, and Raising (Lowering) of Indices

Definition 6.2.8 (Contraction). Let $\left\{V_{i}\right\}_{i \in I}$ be a finite collection of finite $k$-vector spaces such that for some $k, j \in I$ we have $V_{k}=V$ and $V_{j}=V^{*}$. We define the contraction of the tensor product $\bigotimes_{i \in I} V_{i}$ as the linear mapping

$$
\bigotimes_{i \in I} V_{i} \longrightarrow \bigotimes_{\substack{i \in I \\ i \neq i, j}} V_{i} \text { mapping } \otimes_{i \in I} v_{i} \longmapsto v_{j}^{*}\left(v_{k}\right)\left(\otimes_{\substack{i \in I, i \neq j, k}} v_{i}\right)
$$

where $v_{j}^{*}\left(v_{k}\right)=\delta_{j k}$.
Definition 6.2.9 (Raising and Lowering). Let $\left\{V_{i}\right\}_{i=1}^{p}$ be a finite collection of finite $k$ vector spaces and $g: V_{i} \rightarrow V_{i}^{*}$ be an isomorphism. Then we define the lowering of the index $i$ as the linear morphism

$$
\mathrm{id} \otimes \cdots \otimes g \otimes \cdots \otimes \mathrm{id}: \bigotimes_{i=1}^{p} V_{i} \longrightarrow V_{1} \otimes \cdots \otimes V_{i}^{*} \otimes \cdots \otimes V_{p}
$$

Moreover, the raising of the index $i$ is just defined as the inverse of the above linear morphism.

## Tensor Multiplication Functor

Definition 6.2.10 (Tensor multiplication functor). Let FinVect $_{k}$ be the category of finite $k$-vector spaces together with linear morphisms between them. Given $M \in$ FinVect $_{k}$,
we define the functor of tensor multiplication on $M$ as the mapping of objects $L \stackrel{F}{\mapsto} L \otimes M$ and the mapping of morphisms $f \stackrel{F}{\mapsto} f \otimes \mathrm{id}_{M}$. Hence we have $\mathrm{id}_{L} \mapsto \mathrm{id}_{L} \otimes \mathrm{id}_{M}=\mathrm{id}_{L \otimes M}$ and given composable morphisms $f, g \in \operatorname{Mor}\left(\right.$ FinVect $\left._{k}\right)$ we have $f g \mapsto(f g) \otimes \mathrm{id}_{M}=$ $\left(f \otimes \mathrm{id}_{M}\right)\left(g \otimes \mathrm{id}_{M}\right)$ as wanted.

Proposition 6.2.11 (Exactness). Let $0 \rightarrow V \stackrel{f}{\mapsto} L \stackrel{g}{\rightarrow} W \rightarrow 0$ be a short exact sequence, where $V, L, W \in \operatorname{FinVect}_{k}$. Let $M \in$ FinVect $_{k}$, then following sequence is exact

$$
0 \rightarrow V \otimes M \xrightarrow{f \otimes \mathrm{id}_{M}} L \otimes M \xrightarrow{g \otimes \operatorname{id}_{M}} W \otimes W \rightarrow 0 .
$$

Proof. ( $f \otimes \mathrm{id}_{M}$ is injective) Let $v \otimes m \in V \otimes M$ be a non-zero factorizable tensor. Then in particular $f(v) \neq 0$, since $f$ is injective, then $f(v) \otimes m$ is also non-zero, since $m \neq 0$, hence $\operatorname{ker}\left(f \otimes \mathrm{id}_{M}\right)=0 .\left(g \otimes \mathrm{id}_{M}\right.$ is surjective) Let $w \otimes m \in W \otimes M$ be any factorizable tensor, then in particular exists $\ell \in L$ such that $g(\ell)=w$, from the fact that $g$ is surjective. Hence we find that $\ell \otimes m \stackrel{g \otimes \mathrm{id}_{M}}{\longmapsto} g(\ell) \otimes m=w \otimes m$, since the collection of factorizable tensors form a base for the tensor product, we can conclude that $g \otimes \mathrm{id}_{M}$ is surjective. $\left(\operatorname{im}\left(f \otimes \mathrm{id}_{M}\right)=\operatorname{ker}\left(g \otimes \mathrm{id}_{M}\right)\right)$ Suppose $\ell \otimes m \in \operatorname{im}\left(f \otimes \mathrm{id}_{M}\right)$, then in particular we have $\ell \in \operatorname{im} f$ and hence $\ell \in \operatorname{ker} g$ since $\operatorname{im} f=\operatorname{ker} g$. This implies in $\ell \otimes m \in \operatorname{ker}\left(g \otimes \mathrm{id}_{M}\right)$ and hence $\operatorname{im}\left(f \otimes \mathrm{id}_{M}\right) \subseteq \operatorname{ker}\left(g \otimes \mathrm{id}_{M}\right)$. Take now any tensor $\ell^{\prime} \otimes m^{\prime} \in \operatorname{ker}(g \otimes m)$ if $\ell^{\prime} \otimes m^{\prime}=0$ then clearly $\ell^{\prime} \otimes m^{\prime} \in \operatorname{im}\left(f \otimes \mathrm{id}_{M}\right)$, suppose on the contrary that $\ell^{\prime} \otimes m^{\prime} \neq 0$, then certainly $\ell^{\prime} \in \operatorname{ker} g$ and in particular $\ell^{\prime} \in \operatorname{im} f$. Then it follows that $\ell^{\prime} \otimes m^{\prime} \in \operatorname{im}\left(f \otimes \mathrm{id}_{M}\right)$ and hence $\operatorname{ker}\left(g \otimes \mathrm{id}_{M}\right) \subseteq \operatorname{im}\left(f \otimes \mathrm{id}_{M}\right)$. This finishes the proof.

### 6.3 Tensor Algebra

Definition 6.3.1 (Mixed tensor product). Let $V$ be a finite dimensional $k$-vector space. We define the $p$-covariant and $q$-contravariant mixed tensor product on $V$ as

$$
T_{p}^{q}(V)=V^{* \otimes p} \otimes V^{\otimes q}
$$

Elements of such object are called tensors of type $(p, q)$ and rank $p+q$ on $V$. We define $T_{0}^{0}(V)=k$.

Example 6.3.2. The following examples illustrate some specific type of mixed tensor product on $V$, showing that such a construction generalizes various objects in linear algebra.
(a) Tensors of type $(0,0)$ are called scalar tensors of rank 0 .
(b) Tensors of type $(1,0)$ are linear functionals on $V$.
(c) Tensors of type $(0,1)$ are vectors of $V$.
(d) Tensors of type $(1,1)$ are elements of $V^{*} \otimes V \simeq \operatorname{End}_{\text {vect }_{k}}(V)$, that is, linear operators.
(e) Tensors of type $(2,0)$ are elements of $V^{*} \otimes V^{*} \simeq \operatorname{Mor}\left(V^{* *}, V^{*}\right) \simeq \operatorname{Mor}\left(V, V^{*}\right)$ for a finite dimensional $V$. Moreover $V^{*} \otimes V^{*} \simeq(V \otimes V)^{*} \simeq \operatorname{Mor}(V, V ; k)$ of multilinear maps $V \times V \rightarrow k$, that is, the mixed tensors of type $(2,0)$ are inner products.

Definition 6.3.3 (Mixed tensor multiplication). Let $V$ be a finite $k$-vector space, then

$$
T_{p}^{q}(V)=V^{* \otimes p} \otimes V^{\otimes q} \simeq\left(V^{\otimes p} \otimes V^{* \otimes q}\right)^{*},
$$

which in turn is isomorphic to the space of multilinear maps $V^{p} \times V^{* q} \rightarrow k$. Let $f: V^{p} \times V^{* q} \rightarrow k$ and $g: V^{p^{\prime}} \times v^{* q^{\prime}} \rightarrow k$ be multilinear maps, then we define their tensor multiplication as

$$
f \otimes g: V^{p+p^{\prime}} \times V^{*\left(q+q^{\prime}\right)} \rightarrow k
$$

mapping $\left(v_{1}, \ldots, v_{p}, v_{1}^{\prime}, \ldots, v_{p^{\prime}}^{\prime}, u_{1}^{*}, \ldots, u_{p}^{*}, u_{1}^{\prime *}, \ldots, u_{q^{\prime}}^{\prime *}\right)$ into

$$
f\left(v_{1}, \ldots, v_{p}, u_{1}^{*}, \ldots, u_{p}^{*}\right) g\left(v_{1}^{\prime}, \ldots, v_{p^{\prime}}^{\prime}, u_{1}^{* *}, \ldots, u_{q^{\prime}}^{\prime *}\right),
$$

where $v_{i}, v_{i}^{\prime} \in V$ and $u_{j}^{*}, u_{j}^{\prime *} \in V^{*}$. This shows that clearly this tensor multiplication is, in general, non-commutative. However, it is

- (Bilinear) For all $a, b \in k$, then

$$
\left(a f_{1}+b f_{2}\right) \otimes g=a\left(f_{1} \otimes g\right)+b\left(f_{2} \otimes g\right) \text { and } f \otimes\left(a g_{1}+b g_{2}\right)=a\left(f \otimes g_{1}\right)+b\left(f \otimes g_{2}\right)
$$

- (Associative) $(f \otimes g) \otimes h=f \otimes(g \otimes h)$.

Lemma 6.3.4. Let $V$ be a $k$-vector space. There exists a natural isomorphism

$$
T_{p+1}^{q}(V) \simeq\left\{\text { multilinear maps }\left(V^{*}\right)^{p} \times V^{q} \rightarrow V\right\}
$$

Definition 6.3.5 (Tensor algebra). We define the tensor algebra of the $k$-vector space $V$ to be the infinite dimensional $k$-vector space

$$
T(V)=\bigotimes_{p, q=1}^{\infty} T_{p}^{q}(V)
$$

### 6.4 Symmetric Tensors and Symmetric Algebra

Definition 6.4.1 (Symmetric tensor). Let $V$ be a $k$-vector space and consider the mixed tensor product $T_{0}^{q}(V)$. For every permutation $\sigma \in S_{q}$ (where $S_{q}$ denotes the symmetry group of $q$ elements), define the linear transformation

$$
f_{\sigma}: T_{0}^{q}(V) \rightarrow T_{0}^{q}(V), \otimes_{i=1}^{q} v_{i} \stackrel{f_{\sigma}}{\longmapsto} \otimes_{i=1}^{q} v_{f_{\sigma}(i)} .
$$

We say that a tensor $T \in T_{0}^{q}(V)$ is symmetric if for all $\sigma \in S_{q}$

$$
f_{\sigma}(T)=T .
$$

Definition 6.4.2 (Symmetric tensor space). We denote the subspace of $T_{0}^{q}(V)$ consisting of symmetric tensors as the symmetric space $\operatorname{Sym}^{q}(V)$, also called symmetric power.

Definition 6.4.3 (Symmetrization map). Let $V$ be a $k$-vector space. We define the projection operator $S: T_{0}^{q}(V) \rightarrow \operatorname{Sym}^{q}(V)$, which maps factorizable tensors to symmetric tensors.

Proposition 6.4.4 (Symmetric space universal property). Let $V$ be a $k$-vector space. Given any $k$-vector space $L$ and a multilinear symmetric map $\mu$ : $V^{q} \rightarrow L$, there exists a unique linear map $\ell: \operatorname{Sym}^{q}(V) \rightarrow L$ such that $\ell \circ S \circ \otimes=\mu$. That is, the following diagram commutes


Proof. Since $\operatorname{Sym}^{q}(V)$ is a subspace of $T_{0}^{q}(V)$, then we can use Theorem 6.1.4.
Proposition 6.4.5. For char $k \nmid q!$, we have that

$$
S(T)=\frac{1}{q!} \sum_{\sigma \in S_{q}} f_{\sigma}(T)
$$

and hence $S^{2}=S$, and $S\left(T_{0}^{q}(V)\right)=\operatorname{Sym}^{q}(V)$.
Proof. Let $\left\{e_{i}\right\}_{i=1}^{n}$ be a basis for $V$. First, notice that if $T=T^{i_{1}, \ldots, i_{q}} \in \operatorname{Sym}^{q}(V)$, then

$$
S(T)=\frac{1}{q!} \sum_{\sigma \in S_{q}} f_{\sigma}\left(T^{i_{1}, \ldots, i_{q}}\right)=\frac{1}{q!}\left(q!T^{i_{1}, \ldots, i_{q}}\right)=T
$$

where we assumed that char $k \nmid q$ ! in order to obtain a non-zero tensor after the summation. Hence $\left.S\right|_{\operatorname{Sym}^{q}(T)}=\operatorname{id}_{\text {Sym }^{q}(V)}$ therefore we find that $\operatorname{Sym}^{q}(V) \subseteq \mathrm{im} S$. Moreover, let $T=T^{i_{1}, \ldots, i_{q}} \in T_{0}^{q}(V)$ be any factorizable tensor, then its image under $S$ is clearly a symmetric tensor, that is $S(T) \in \operatorname{Sym}^{q}(V)$, hence $\operatorname{im} S \subseteq \operatorname{Sym}^{q}(V)$. This implies that $\operatorname{im} S=\operatorname{Sym}^{q}(V)$ for char $k \nmid q$ !. Moreover, $S(S(T))=S(T)$ from the first argument, since $S(T) \in \operatorname{Sym}^{q}(V)$.

Remark 6.4.6. From now on we are going to assume that char $k \nmid q$ !
Notation 6.4.7. Notice that since $\operatorname{im} S=\operatorname{Sym}^{q}(V)$ then we can write any symmetrized tensor in any permutation that we choose, hence, we sometimes use the dot notation $\operatorname{Sym}^{q}(V) \ni v_{1} \otimes \cdots \otimes v_{q}=v_{1} \cdot \ldots \cdot v_{q}$. Moreover, if $\left\{e_{i}\right\}_{i=1}^{n}$ is a basis for $V$, one can even adopt an exponential notation $e_{1} \cdot \ldots \cdot e_{q}=e_{1}^{a_{1}} \cdot \ldots \cdot e_{q}^{a_{n}}$, where $a_{i}$ denotes the number of times the component $e_{i}$ appears in the factorizable tensor, and $a_{1}+\cdots+a_{q}=q$.

Proposition 6.4.8 (Basis for $\operatorname{Sym}^{q}(V)$ ). Let $\left\{e_{i}\right\}_{i=1}^{n}$ be a basis for the $k$-vector space $V$. Then, the collection of tensors $\left\{e_{1}^{a_{1}} \cdot \ldots \cdot e_{n}^{a_{n}}: a_{1}+\cdots+a_{n}=q\right\} \subseteq \operatorname{Sym}^{q}(V)$ form a basis for the symmetric space $\operatorname{Sym}^{q}(V)$. This implies that $\operatorname{Sym}^{q}(V)$ is the subspace of $k\left[e_{1}, \ldots, e_{n}\right]$ of homogeneous polynomials of total degree $q$.

Proof. We know from Lemma 6.1.8 that $\left\{\otimes_{j=1}^{q} e_{i_{j}}: 1 \leqslant i_{j} \leqslant n\right\}$ forms a basis for $T_{0}^{q}(V)$. Since $S$ is multilinear, then $S\left(\left\{\otimes_{j=1}^{q} e_{i_{j}}\right\}\right)=\operatorname{Sym}^{q}(V)$, that is $\left\{e_{i_{1}}^{a_{1}} \cdot \ldots \cdot e_{n}^{a_{n}}\right\}=\left\{S\left(\otimes_{j=1}^{q} e_{i_{j}}\right)\right\}$ generates the space $\operatorname{Sym}^{q}(V)$. Hence, to show that the collection of symmetric tensors of the basis elements of $V$ forms a basis for the symmetric space, we just need to show their linear independence.

To do that, fix a sequence of indices $I=\left(i_{1}, \ldots, i_{q}\right)$ such that $1 \leqslant i_{1} \leqslant \ldots \leqslant i_{q} \leqslant n$. For all $v_{j}=\sum_{i=1}^{n} a_{i j} e_{i} \in V$ where $1 \leqslant j \leqslant q$, let $\mu_{I}: V^{q} \rightarrow k$ be the mapping

$$
\mu_{I}\left(v_{1}, \ldots, v_{n}\right)=\sum_{\sigma \in S_{q}} \prod_{j=1}^{q} e_{i_{\sigma(j)}}^{*}\left(v_{j}\right)
$$

From the universal property of tensor products we find a unique functional $f_{I}: \operatorname{Sym}^{q}(V) \rightarrow$ $k$ such that the diagram commutes:

hence for all $\left(v_{1}, \ldots, v_{q}\right) \in T_{0}^{q}(V)$ we have $\mu\left(v_{1}, \ldots, v_{q}\right)=f\left(v_{1} \cdot \ldots \cdot v_{n}\right)$. Notice that if we have a monotonically increasing sequence of indices $I^{\prime}=\left(i_{1}^{\prime}, \ldots, i_{q}^{\prime}\right) \neq I$, there must be an index $i_{j_{0}}^{\prime} \in I^{\prime}$ such that $i_{j_{0}}^{\prime} \neq i_{j}$ for all $i_{j} \in I$. This way we find that for all $\sigma \in S_{q}$ the product $\prod_{j=1}^{q} e_{i_{\sigma(j)}}^{*}\left(e_{i_{j}^{\prime}}\right)=0$ because $e_{i_{\sigma(j)}}^{*}\left(e_{i_{j_{0}^{\prime}}}\right)=0$. In particular, this implies in

$$
\mu_{I}\left(e_{i_{1}^{\prime}}, \ldots, e_{i_{q}^{\prime}}\right)=f_{I}\left(e_{i_{1}^{\prime}} \cdot \ldots \cdot e_{i_{q}^{\prime}}\right)=0
$$

Lets compute the value of $f_{I}\left(e_{i_{1}} \cdot \ldots \cdot e_{i_{q}}\right)$. To do this, consider again the fixed monotone increasing sequence of indices that we started with, that is, I. Let $Q \leqslant q$ denote the number of distinct values of $i_{j} \in I$ for $1 \leqslant j \leqslant q$. Define now, for all $1 \leqslant r \leqslant Q$ the index sets $J_{r}:=\left\{j: e_{i_{j}}\right.$ have equal values $\}$. Define $n_{r}:=\left|J_{r}\right|$, this construction yields a total of $\prod_{r=1}^{Q}\left(n_{r}!\right)$ permutations. Notice that all permutations of $J_{r}$ 's are such that $\sigma_{J_{r}}(I)=I$ (they change the position of equal elements, causing no alteration of the sequence), we'll denote such permutations by $\sigma_{J} \in S_{q}$. Moreover, these are all the possible permutations that leave $I$ unaltered. Hence

$$
f_{I}\left(e_{i_{1}} \cdot \ldots \cdot e_{i_{q}}\right)=\mu_{I}\left(e_{i_{1}}, \ldots, e_{i_{q}}\right)=\sum_{\sigma \in S_{q}} \prod_{j=1}^{q} e_{i_{\sigma(j)}}^{*}\left(e_{j}\right)=\sum_{\sigma_{J} \in S_{q}} \prod_{j=1}^{q} e_{i_{\sigma_{J}(j)}}^{*}\left(e_{j}\right)=\prod_{r=1}^{Q}\left(n_{r}!\right) \in k .
$$

Therefore, if char $k \nmid q!$, we find $f_{I}\left(e_{i_{1}} \cdot \ldots \cdot e_{i_{q}}\right) \neq 0 \in k$.

If there exists a linear relation

$$
\sum_{i_{1}^{\prime}, \ldots, i_{q}^{\prime}} c_{i_{1}^{\prime}, \ldots, i_{q}^{\prime}}\left(e_{i_{1}^{\prime}} \cdot \ldots \cdot e_{i_{q}^{\prime}}\right)=0
$$

where the indices are arranged in monotonic increasing order, then we conclude that

$$
\begin{aligned}
0=f_{I}\left(\sum_{i_{1}^{\prime}, \ldots, i_{q}^{\prime}} c_{i_{1}^{\prime}, \ldots, i_{q}^{\prime}}\left(e_{i_{1}^{\prime}} \cdot \ldots \cdot e_{i_{q}}\right)\right) & =c_{i_{1}^{\prime}, \ldots, i_{q}^{\prime}} \sum_{\sigma \in S_{q}} \prod_{j=1}^{q} e_{i_{\sigma(j)}}^{*}\left(e_{i_{j}^{\prime}}\right) \\
& =c_{i_{1}, \ldots, i_{q}} \sum_{\sigma \in S_{q}} \prod_{j=1}^{q} e_{i_{\sigma(j)}}^{*}\left(e_{j}\right) \\
& =c_{i_{1}, \ldots, i_{q}} f_{I}\left(e_{i_{1}} \cdot \ldots \cdot e_{i_{q}}\right)
\end{aligned}
$$

hence $c_{i_{1}, \ldots, i_{q}}=0$ since, as proved above, $f_{I}\left(e_{i_{1}} \cdot \ldots \cdot e_{i_{q}}\right) \neq 0$. Now, passing $I$ through all the monotone increasing sequences of indices, we prove that all of the coefficients $c_{i_{1}^{\prime}, \ldots, i_{c}^{\prime}}$ vanish, as wanted. This proves that $\left\{e_{1}^{a_{1}} \cdot \ldots \cdot e_{n}^{a_{n}}: a_{1}+\cdots+a_{n}=q\right\}$ is indeed a linearly independent collection.

Corollary 6.4.9. If $V$ is a finite $n$-dimensional $k$-vector space, then

$$
\operatorname{dim}_{k}\left(\operatorname{Sym}^{q}(V)\right)=\binom{n+q-1}{q}
$$

Proof. Notice that from Proposition 6.4.8 we have that $\operatorname{Sym}^{q}(V)$ is the subspace of $k\left[e_{1}, \ldots, e_{n}\right]$ consisting of homogeneous polynomials of degree $q$. In order to find the dimension of such space, it is sufficient to find the cardinality $\left|\left\{a_{1}+\cdots+a_{n}=q: a_{i} \geqslant 0\right\}\right|$. A combinatorial way to do so is to define $q$ symbols | (which represent the unity 1 ), which can be divided into $n$ different groups. In order to do so, we introduce $n-1$ symbols (we are going to use +), which are arranged in order to separate the bars into the $n$ requested groups (example: for $n=3$ and $q=5$, one possible arrangement is $\mid+\|+\|)$. We have a total of $n+q-1$ symbols, and we want $q$ of those (the bars) in order to create the groupings, that is, we have a total of $\binom{n+q-1}{q}$ ways of doing so. This proves the corollary.

Definition 6.4.10. We define the symmetric algebra over a finite $n$-dimensional $k$-linear space $V$ to be the object $\operatorname{Sym} V=\bigoplus_{q=1}^{\infty} \operatorname{Sym}^{q} V$. This is the space of all polynomials in $e_{1}, \ldots, e_{n}$ variables (where $e_{1}, \ldots, e_{n}$ form a basis for $V$ ).

We define multiplication in Sym $V$ as the map

$$
\begin{equation*}
\operatorname{Sym}^{q} V \times \operatorname{Sym}^{d} V \rightarrow \operatorname{Sym}^{q+d} V, \text { mapping }\left(T_{1}, T_{2}\right) \mapsto S\left(T_{1} \otimes T_{2}\right)=: T_{1} T_{2} \tag{6.1}
\end{equation*}
$$

Proposition 6.4.11. The multiplicative structure defined by Eq. (6.1) makes Sym $V$ a commutative associative $k$-algebra. Under the isomorphism of tensors and homogeneous polynomials as described in Proposition 6.4.8, this multiplicative structure
preserves that of the multiplication of polynomials, making such isomorphism an isomorphism of algebras.

$$
\left\{\text { polynomials in } e_{1}, \ldots, e_{n}\right\} \simeq \operatorname{Sym} V
$$

where $e_{1}, \ldots, e_{n}$ forms a basis for $V$.
Proof. Let $T_{1} \in T_{0}^{q}(V)$ and $T_{2} \in T_{0}^{d}(V)$, then we find that

$$
S\left(T_{1}\right) \otimes T_{2}=\left(\frac{1}{q!} \sum_{\sigma \in S_{q}} f_{\sigma}\left(T_{1}\right)\right) \otimes T_{2}
$$

This way we find that

$$
\begin{aligned}
S\left(S\left(T_{1}\right) \otimes T_{2}\right)=S\left(\frac{1}{q!} \sum_{\sigma \in S_{q}} f_{\sigma}\left(T_{1}\right) \otimes T_{2}\right) & =\frac{1}{q!} \sum_{\sigma \in S_{q}} S\left(f_{\sigma}\left(T_{1}\right) \otimes T_{2}\right) \\
& =\frac{1}{q!} \sum_{\sigma \in S_{q}} S\left(T_{1} \otimes T_{2}\right) \\
& =S\left(T_{1} \otimes T_{2}\right)
\end{aligned}
$$

where we used the fact that $S\left(f_{\sigma}\left(T_{1}\right) \otimes T_{2}\right)=S\left(T_{1} \otimes T_{2}\right)$. Hence, we can conclude in general that

$$
S\left(S\left(T_{1}\right) \otimes T_{2}\right)=S\left(T_{1} \otimes S\left(T_{2}\right)\right)=S\left(T_{1} \otimes T_{2}\right)
$$

If we extend such argument for tensors $T_{1}, T_{2}, T_{3}$, we find that

$$
\left(T_{1} T_{2}\right) T_{3}=S\left(S\left(T_{1} \otimes T_{2}\right) \otimes T_{3}\right)=T_{1}\left(T_{2} T_{3}\right)=S\left(T_{1} \otimes\left(T_{1} \otimes T_{3}\right)\right)=T_{1} T_{2} T_{3}=S\left(T_{1} \otimes T_{2} \otimes T_{3}\right)
$$

which proves associativity for the multiplication. From tensor commutativity we find that

$$
T_{1} T_{2}=S\left(T_{1} \otimes T_{2}\right)=S\left(T_{2} \otimes T_{1}\right)=T_{2} T_{1}
$$

Finally, notice that from this multiplicative structure we have that the multiplication of polynomials is given by $\left(e_{1}^{a_{1}} \cdots e_{n}^{a_{n}}\right)\left(e_{n}^{b_{1}} \cdots e_{n}^{b_{n}}\right)=e_{1}^{a_{1}+b_{1}} \cdots e_{n}^{a_{n}+b_{n}}$, which shows the algebra isomorphism.

### 6.5 Alternating Tensors and Exterior Powers

Definition 6.5.1 (Alternating map). Let $V$ and $W$ be vector spaces. A multilinear map $f: V^{n} \rightarrow W$ is called alternating if for any $\left(v_{1}, \ldots, v_{n}\right) \in V^{n}$ such that exists $i<j$ for which $v_{i}=v_{j}$ then

$$
f\left(v_{1}, \ldots, v_{n}\right)=0 .
$$

Proposition 6.5.2. Let $f: V^{n} \rightarrow W$ be an alternating multilinear map. Then if $\tau \in S_{n}$ is a transposition, we have

$$
f\left(v_{\tau(1)}, \ldots, v_{\tau(n)}\right)=-f\left(v_{1}, \ldots, v_{n}\right)
$$

In general, if $\sigma \in S_{n}$ is any permutation, then

$$
f\left(v_{\sigma(1)}, \ldots, v_{\sigma(n)}\right)=\operatorname{sign}(\sigma) f\left(v_{1}, \ldots, v_{n}\right)
$$

Proof. Let $\tau$ be a transposition between indices $i<j$, then consider

$$
\begin{aligned}
0=f\left(v_{1}, \ldots, v_{i}+v_{j}, \ldots, v_{i}+v_{j}, \ldots, v_{n}\right)= & f\left(v_{1}, \ldots, v_{i}, \ldots, v_{i}+v_{j}, \ldots, v_{n}\right) \\
& +f\left(v_{1}, \ldots, v_{j}, \ldots, v_{i}+v_{j}, \ldots, v_{n}\right)
\end{aligned}
$$

therefore we find

$$
\begin{aligned}
f\left(v_{1}, \ldots, v_{i}, \ldots, v_{i}+v_{j}, \ldots, v_{n}\right) & =-f\left(v_{1}, \ldots, v_{j}, \ldots, v_{i}+v_{j}, \ldots, v_{n}\right) \\
f\left(v_{1}, \ldots, v_{i}, \ldots, v_{j}, \ldots, v_{n}\right) & =-f\left(v_{1}, \ldots, v_{j}, \ldots, v_{i}, \ldots, v_{n}\right) \\
f\left(v_{1}, \ldots, v_{n}\right) & =f\left(v_{\tau(1)}, \ldots, v_{\tau(n)}\right)
\end{aligned}
$$

where we used that $f\left(v_{1}, \ldots, v_{i}, \ldots, v_{i}, \ldots, v_{n}\right)=f\left(v_{1}, \ldots, v_{j}, \ldots, v_{j}, \ldots, v_{n}\right)=0$. This shows the first proposition. For the second proposition, by means of Proposition 7.2.14 and Proposition 7.2.8 and the above proposition for transpositions, we conclude the proof.

Remark 6.5.3. Notice that Proposition 6.5.2 is not sufficient to characterize alternating maps, it is but a necessary property. A counterexample to that is the field with characteristic 2.

## Alternating Tensor and Exterior Powers

Definition 6.5.4 (Alternating tensor). Let $V$ be a $k$-vector space we define a tensor $T \in T_{0}^{q}(V)$ to be alternating if for all $\sigma \in S_{q}$ permutation we have

$$
f_{\sigma}(T)=\operatorname{sign}(\sigma) T
$$

We denote by $\Lambda^{q} V$ the subspace of $T_{0}^{q}(V)$ of alternating tensors and call it the $q$-exterior power of $V$.

Proposition 6.5.5 (Alternating tensor projection). Let $V$ be a $k$-vector space and char $k \nmid$ $q$ !. We define the linear projection operator $A: T_{0}^{q}(V) \rightarrow T_{0}^{q}(V)$ where

$$
A(T)=\frac{1}{q!} \sum_{\sigma \in S_{q}} \operatorname{sign}(\sigma) f_{\sigma}(T)
$$

Then $A^{2}=A$ and $\operatorname{im} A=\Lambda^{q} V$.

Proof. First we show that $\operatorname{im} A \subseteq \Lambda^{q} V$ (we already know that clearly $\Lambda^{q} V \subseteq \operatorname{im} A$ ). Notice that

$$
\begin{aligned}
f_{\sigma}(A(T)) & =f_{\sigma}\left(\frac{1}{q!} \sum_{\tau \in S_{q}} \operatorname{sign}(\tau) f_{\tau}(T)\right) \\
& =\frac{1}{q!} \sum_{\tau \in S_{q}} \operatorname{sign}(\tau) f_{\sigma \tau}(T) \\
& =\operatorname{sign}(\sigma)\left(\frac{1}{q!} \sum_{\tau \in \mathcal{S}_{q}} \operatorname{sign}(\sigma \tau) f_{\sigma \tau}(T)\right) \\
& =\operatorname{sign}(\sigma) A(T)
\end{aligned}
$$

where we've used Proposition 7.2.14. Hence we conclude that $\operatorname{im} A=\Lambda^{q} V$. For the second proposition, notice that

$$
A^{2}=\frac{1}{q!^{2}} \sum_{\sigma, \tau \in S_{q}} \operatorname{sign}(\sigma \tau) f_{\sigma \tau}=\frac{1}{q!} \sum_{\rho \in S_{q}} \operatorname{sign}(\rho) f_{\rho}=A
$$

since any permutation $\rho \in S_{q}$ can be represented in $q$ ! different ways as a form of a product $\sigma \tau$. This concludes the proof.

Definition 6.5.6 (Exterior multiplication). Let $V$ be a $k$-vector space and $v_{1} \otimes \cdots \otimes v_{q} \in$ $T_{0}^{q}(V)$. We define the exterior multiplication to be the map

$$
A\left(v_{1} \otimes \cdots \otimes v_{q}\right):=v_{1} \wedge \cdots \wedge v_{q}
$$

Proposition 6.5.7 (Universal property for exterior power). Let $V$ be a $k$-vector space. For all $k$-vector spaces $L$ together with an alternating multilinear map $\mu$ : $V^{q} \rightarrow L$, there exists a unique $k$-linear map $\ell: \Lambda^{q} V \rightarrow L$ for which the diagram commutes


Proof. Since $\Lambda^{q} V$ is a subspace of $T_{0}^{q}(V)$, then we can use the universal property Theorem 6.1.4.

Now we get some geometric motivation behind the algebraic structure of the exterior power $\Lambda^{q} V$. Alternating tensors $v_{1} \wedge \cdots \wedge v_{q} \in \Lambda^{q} V$ can be seen as $q$-dimensional oriented volume elements, where by oriented we mean that the transposition of two edges, say $v_{j}$ and $v_{i}$, implies in a change up to a minus sign of the value (the sign is what creates the bridge between orientation and wedge product). Notice that when there are equal edges, the volume becomes malformed and $v_{1} \wedge \cdots \wedge v_{q}=0$.

Proposition 6.5.8. Let $V$ be a $n$-dimensional $k$-vector space, where char $k \neq 2$. Define $\left\{e_{i}\right\}_{i=1}^{n}$ to be a base for $V$. Then the exterior product $e_{i_{1}} \wedge \cdots \wedge e_{i_{q}}=0$ if there exists $i_{a}=i_{b}$ for some $1 \leqslant a, b \leqslant q$.

Proof. Since $A$ is an alternating map, we can use Definition 6.5.1 and conclude the proof.

For the next proposition, we proceed in a similar fashion as in Proposition 6.4.8.
Proposition 6.5.9 (Exterior power $\Lambda^{q} V$ basis). Let $\left\{e_{i}\right\}_{i=1}^{n}$ be a basis for the $k$-vector space $V$. The factorizable tensors

$$
A\left(e_{i_{1}} \otimes \cdots \otimes e_{i_{q}}\right)=e_{i_{1}} \wedge \cdots \wedge e_{i_{q}}
$$

with $1 \leqslant i_{1}<\cdots<i_{q} \leqslant n$ form a basis for the subspace $\Lambda^{q} V$.
Proof. Let $B:=\left\{e_{i_{1}} \wedge \cdots \wedge e_{i_{q}}: 1 \leqslant i_{1}<\cdots<i_{q} \leqslant n\right\}$. Since $\left\{e_{i_{1}} \otimes \cdots \otimes e_{i_{q}}: 1 \leqslant\right.$ $\left.i_{1}<\ldots i_{q} \leqslant n\right\}$ generates $T_{0}^{q}(V)$ (see Lemma 6.1.8), then clearly $B$ does generate the subspace $\Lambda^{q} V$. We now show that $B$ is linearly independent.

Denote by $\mathcal{I}:=\left\{I=\left(i_{j}\right)_{j=1}^{q}: 1 \leqslant i_{1}<\cdots<i_{q} \leqslant n\right\}$ the set of strictly increasing $q$-tuples. For each $I:=\left(i_{1}, \ldots, i_{q}\right) \in \mathcal{I}$ we define the alternating multilinear map $\mu_{I}: V^{q} \rightarrow k$ as

$$
\mu_{I}\left(v_{1}, \ldots, v_{q}\right)=\sum_{\sigma \in S_{q}} \operatorname{sign}(\sigma) \prod_{j=1}^{q} e_{i_{\sigma(j)}}^{*}\left(v_{i_{j}}\right)
$$

where $\left\{e_{i_{j}}^{*}\right\}_{i_{j} \in I}$ is the dual of $\left\{e_{i_{j}}\right\}_{i_{j} \in I}$. Using Proposition 6.5 .7 we can define a unique linear map $f_{I}: \Lambda^{q} V \rightarrow k$ such that the diagram commutes

which implies in $\mu_{I}\left(v_{1}, \ldots, v_{q}\right)=f_{I}\left(v_{1} \wedge \cdots \wedge v_{q}\right)$ for all $\left(v_{1}, \ldots, v_{q}\right) \in V^{q}$.
Let $I^{\prime}=\left(i_{1}^{\prime}, \ldots, i_{q}^{\prime}\right) \in I$ such that $I^{\prime} \neq I$. From the strictly ordering of the indices, we conclude that there must exists some $1 \leqslant j_{0} \leqslant q$ such that $i_{j_{0}}^{\prime} \neq i_{j}$ for all $1 \leqslant j \leqslant q$. In particular, this implies that

$$
f_{I}\left(e_{i_{1}^{\prime}} \wedge \cdots \wedge e_{i_{q}^{\prime}}\right)=\mu_{I}\left(e_{i_{1}^{\prime}}, \ldots, e_{i_{q}^{\prime}}\right)=\sum_{\sigma \in S_{q}} \operatorname{sign}(\sigma) \prod_{j=1}^{q} e_{i_{\sigma(j)}}^{*}\left(e_{i_{j}^{\prime}}\right)=0
$$

since $e_{i_{\sigma(j)}}^{*}\left(e_{i_{0}}\right)=0$ for all permutations $\sigma \in S_{q}$ and $1 \leqslant j \leqslant q$. On the other hand, we have that
$f_{I}\left(e_{i_{1}} \wedge \cdots \wedge e_{i_{q}}\right)=\mu_{I}\left(e_{i_{1}}, \ldots, e_{i_{q}}\right)=\sum_{\sigma \in S_{q}} \operatorname{sign}(\sigma) \prod_{j=1}^{q} e_{i_{\sigma(j)}}^{*}\left(e_{i_{j}}\right)=\operatorname{sign}(\mathrm{id}) \prod_{j=1}^{q} e_{i_{\mathrm{id}(j)}}^{*}\left(e_{i_{j}}\right)=1$

Let $c_{P} \in k$ for all $P \in I$ and consider the vanishing linear combination

$$
\begin{equation*}
\sum_{P \in I} c_{P}\left(e_{p_{1}} \wedge \cdots \wedge e_{p_{q}}\right)=0 \tag{6.2}
\end{equation*}
$$

Then, if we look at its image under the map $f_{I}$ for all $I \in I$ we conclude that

$$
0=f_{I}\left(\sum_{P \in I} c_{P}\left(e_{p_{1}} \wedge \cdots \wedge e_{p_{q}}\right)\right)=\sum_{P \in I} c_{P} f_{I}\left(e_{p_{1}} \wedge \cdots \wedge e_{p_{q}}\right)=c_{I}
$$

hence we conclude that the linear combination Eq. (6.2) vanishes if and only if each coefficient vanishes. This concludes that $B$ is linearly independent. Hence we've proved that $B$ is a basis for $\Lambda^{q} V$.

Proposition 6.5.10. Let $V$ be a $n$-dimensional $k$-vector space. Then we have

$$
\operatorname{dim}_{k} \Lambda^{q} V=\binom{n}{q}
$$

Proof. Since we want our indices to be strictly increasing, we are left with $n$ elements of which we want to arrange in groups of $q$ elements. Hence the number of possible arrangements is $\binom{n}{q}$ and this concludes the proof.

## Exterior Algebra

Definition 6.5.11 (Exterior algebra). Let $V$ be a $k$-vector space. We define the exterior algebra on $V$ as

$$
\Lambda^{\bullet} V=\bigoplus_{q=0}^{\infty} \Lambda^{q} V
$$

together with a multiplicative structure $\wedge: \Lambda^{d} V \otimes \Lambda^{q} V \rightarrow \Lambda^{d+q} V$ mapping

$$
\left(v_{1} \wedge \cdots \wedge v_{d}\right) \otimes\left(w_{1} \wedge \cdots \wedge w_{q}\right) \stackrel{\wedge}{\mapsto} v_{1} \wedge \cdots \wedge v_{d} \wedge w_{1} \wedge \cdots \wedge w_{q}
$$

We interpret $\Lambda^{0} V=k$.
Proposition 6.5.12. The multiplicative structure of the tensor algebra $\Lambda^{\bullet} V$ is associative and skew-commutative, that is, for all $\alpha \in \Lambda^{d} V$ and $\beta \in \Lambda^{q} V$ we have $\alpha \wedge \beta=(-1)^{d q} \beta \wedge \alpha$.

Proof. First, we prove that if $T_{1} \in T_{0}^{q}(V)$ and $T_{2} \in T_{0}^{d}(V)$ then

$$
\begin{equation*}
A\left(A\left(T_{1}\right) \otimes T_{2}\right)=A\left(T_{1} \otimes A\left(T_{2}\right)\right)=A\left(T_{1} \otimes T_{2}\right) \tag{6.3}
\end{equation*}
$$

Notice that

$$
A\left(T_{1}\right) \otimes T_{2}=\frac{1}{q!} \sum_{\sigma \in S_{q}} \operatorname{sign}(\sigma) f_{\sigma}\left(T_{1}\right) \otimes T_{2}
$$

therefore we find

$$
\begin{equation*}
A\left(A\left(T_{1}\right) \otimes T_{2}\right)=\frac{1}{q!} \sum_{\sigma \in S_{q}} \operatorname{sign}(\sigma) A\left(f_{\sigma}\left(T_{1}\right) \otimes T_{2}\right) \tag{6.4}
\end{equation*}
$$

Moreover, we can construct an injection of the symmetry groups $S_{q} \rightarrow S_{q+d}$ via the mapping $\sigma \mapsto \bar{\sigma}$ where $\bar{\sigma}(i)=\sigma(i)$ for all $i \in\{1, \ldots, q\}$ and $\bar{\sigma}(i)=i$ for all $i \in\{q+1, \ldots, q+d\}$. In particular, we find that $f_{\sigma}\left(T_{1}\right) \otimes T_{2}=f_{\bar{\sigma}}\left(T_{1} \otimes T_{2}\right)$ and clearly $\operatorname{sign}(\bar{\sigma})=\operatorname{sign}(\sigma)$. This implies that
$A\left(f_{\sigma}\left(T_{1}\right) \otimes T_{2}\right)=A\left(f_{\bar{\sigma}}\left(T_{1} \otimes T_{2}\right)\right)=f_{\bar{\sigma}}\left(A\left(T_{1} \otimes T_{2}\right)\right)=\operatorname{sign}(\bar{\sigma}) A\left(T_{1} \otimes T_{2}\right)=\operatorname{sign}(\sigma) A\left(T_{1} \otimes T_{2}\right)$
If we substitute Eq. (6.5) in Eq. (6.4) we find

$$
A\left(A\left(T_{1}\right) \otimes T_{2}\right)=\frac{1}{q!} \sum_{\sigma \in S_{q}} \operatorname{sign}^{2}(\sigma) A\left(T_{1} \otimes T_{2}\right)=A\left(T_{1} \otimes T_{2}\right)
$$

In the same manner we can show that $A\left(T_{1} \otimes A\left(T_{2}\right)\right)=A\left(T_{1} \otimes T_{2}\right)$. Hence Eq. (6.3) holds.
Let $\alpha \in \Lambda^{d} V, \beta \in \Lambda^{q} V, \gamma \in \Lambda^{p} V$. Notice that from the construction of the product map $\wedge$ and Eq. (6.3) we find

$$
(\alpha \wedge \beta) \wedge \gamma=\wedge(\wedge(\alpha \otimes \beta) \otimes \gamma)=\wedge(\alpha \otimes \wedge(\beta \otimes \gamma))=(\alpha \wedge \beta) \wedge \gamma
$$

which proves associativity of the tensor algebra.
We now prove skew-commutativity. Let $\alpha \in \Lambda^{d} V$ and $\beta \in \Lambda^{q} V$, and a permutation $\sigma \in S_{q+d}$ such that $\sigma(i)=q+d-i+1$, consisting of $d q$ transpositions. Hence we find that

$$
\beta \wedge \alpha=f_{\sigma}(\alpha \wedge \beta)=\operatorname{sign}(\sigma)(\alpha \wedge \beta)=(-1)^{d q}(\alpha \wedge \beta)
$$

Proposition 6.5.13. Let $V$ be a $k$-vector space and $\left\{v_{i}\right\}_{i=1}^{d} \subseteq V$. The vectors $v_{1}, \ldots, v_{d}$ are linearly independent if and only if

$$
\Lambda^{d} V \ni v_{1} \wedge \cdots \wedge v_{d} \neq 0
$$

Proof. Let $\left\{v_{i}\right\}_{i=1}^{d}$ be a set of linearly dependent vectors and choose a set of not-all zero scalars $\left\{c_{i}\right\}_{i=1}^{d} \subseteq k$ such that $\sum_{i=1}^{d} c_{i} v_{i}=0$. Choose $1 \leqslant j \leqslant d$ such that $c_{j} \neq 0$ and write $v_{j}=-\sum_{i=1}^{d} \frac{c_{i}}{c_{j}} v_{i}$. Then we find that (recall $\mathcal{M}_{0}$ from our construction of the tensor
product)

$$
\begin{align*}
v_{1} \wedge \cdots \wedge v_{j} \wedge \cdots \wedge v_{d} & =v_{1} \wedge \cdots \wedge\left(-\sum_{i=1}^{d} \frac{c_{i}}{c_{j}} v_{i}\right) \wedge \cdots \wedge v_{d} \\
& =\left(v_{1}, \ldots,-\sum_{i=1}^{d} \frac{c_{i}}{c_{j}} v_{i}, \ldots, v_{d}\right)+\mathcal{M}_{0} \\
& =-\sum_{i=1}^{d} \frac{c_{i}}{c_{j}}\left(v_{1}, \ldots, v_{i}, \ldots, v_{d}\right)+\mathcal{M}_{0}  \tag{6.6}\\
& =-\sum_{i=1}^{d} \frac{c_{i}}{c_{j}}\left(v_{1} \wedge \cdots \wedge v_{i} \wedge \ldots v_{d}\right)  \tag{6.7}\\
& =0 \tag{6.8}
\end{align*}
$$

Where Eq. (6.8) comes from the fact that we have a repeated $v_{i}$ in the wedge product Eq. (6.7).

Suppose $\left\{v_{i}\right\}_{i=1}^{d}$ is a linearly independent set. Then from Proposition 5.6.16 we can build $B \supseteq\left\{v_{i}\right\}_{i=0}^{d}$ such that $B$ is a basis for $V$. From Proposition 6.5.9, we find that $\mathcal{B}=\left\{v_{i_{1}} \wedge \cdots \wedge v_{i_{d}}: v_{i_{j}} \in B\right.$ and $\left.i_{1}<\cdots<i_{d}\right\}$ is a basis for $\Lambda^{d} V$. In particular, $v_{1} \wedge \cdots \wedge v_{d} \in \mathcal{B}$, hence necessarily $v_{1} \wedge \cdots \wedge v_{d} \neq 0$.

Therefore one can trivially see that the kernel of the alternating projection $A$ is simply the collection of all decomposable tensors $v_{1} \otimes \cdots \otimes v_{d}$ such that the collection $\left\{v_{1}, \ldots, v_{d}\right\} \subseteq V$ is linearly dependent.

Corollary 6.5.14. The kernel of the alternating projection $A: T_{0}^{d}(V) \rightarrow \Lambda^{d} V$ is given by the collection $\left\{v_{1} \otimes \cdots \otimes v_{d} \in T_{0}^{d}(V)\right\}$ such that $\left\{v_{1} \otimes \cdots \otimes v_{d}\right\}$ is linearly dependent on $V$.

Theorem 6.5.15. Let $V$ be a finite dimensional $k$-vector space. The map $\phi: \Lambda^{d} V^{*} \xlongequal{\simeq}$ $\left(\Lambda^{d} V\right)^{*}$ given by the mapping

$$
f_{1} \wedge \cdots \wedge f_{d} \stackrel{\phi}{\longmapsto}\left(v_{1} \wedge \cdots \wedge v_{d} \mapsto \sum_{\sigma \in S_{d}} \operatorname{sign}(\sigma) \prod_{j=1}^{d} f_{j}\left(v_{\sigma(j)}\right)\right)
$$

is a $k$-linear isomorphism.

Proof. First of all, we show that $\phi$ is a $k$-linear map. For each $f \in \Lambda^{d} V^{*}$, define the $\operatorname{map} \psi_{f} \in\left(\Lambda^{d} V\right)^{*}$ for which $\phi(f)=\psi_{f}$. Notice that if $a \in k$ and $f, g \in \Lambda^{d}\left(V^{*}\right)$, then
$\phi(f+a g)=\psi_{f+a g}$. Notice that

$$
\begin{aligned}
\psi_{f+a g}\left(v_{1} \wedge \cdots \wedge v_{d}\right) & =\sum_{\sigma \in S_{d}} \operatorname{sign}(\sigma) \prod_{j=1}^{d}\left(f_{j}+a g_{j}\right)\left(v_{\sigma(j)}\right) \\
& =\sum_{\sigma \in S_{d}} \operatorname{sign}(\sigma) \prod_{j=1}^{d} f_{j}\left(v_{\sigma(j)}\right)+a g_{j}\left(v_{\sigma(j)}\right) \\
& =\sum_{\sigma \in S_{d}} \operatorname{sign}(\sigma) \prod_{j=1}^{d} f_{j}\left(v_{\sigma(j)}\right)+a \sum_{\sigma \in S_{d}} \operatorname{sign}(\sigma) \prod_{j=1}^{d} g_{j}\left(v_{\sigma(j)}\right) \\
& =\psi_{f}+a \psi_{g}
\end{aligned}
$$

Thus $\phi(f+a g)=\phi(f)+a \phi(g)$ and therefore $\phi$ is indeed linear.
Suppose $\operatorname{dim} V=n$ and let $\left\{e_{j}\right\}_{j=1}^{n}$ be a base for $V$. Define the collection of strictly increasing $d$-tuples $I=\left\{I=\left(i_{j}\right)_{j=1}^{d}: 1 \leqslant i_{1}<\cdots<i_{d} \leqslant n\right\}$. From Proposition 6.5.9, we know that $\left\{e_{i_{1}} \wedge \cdots \wedge e_{i_{d}} \in \Lambda^{d} V: i_{j} \in I\right.$ and $\left.I \in \mathcal{I}\right\}$ is a basis for $\Lambda^{d} V$. Moreover, $\left\{e_{j}^{*}\right\}_{j=1}^{n}$ is a basis for the dual space $V^{*}$, thus the collection $\left\{e_{i_{1}}^{*} \wedge \cdots \wedge e_{i_{d}}^{*} \in \Lambda^{d} V: i_{j} \in I\right.$ and $\left.I \in \mathcal{I}\right\}$ is a basis for $\Lambda^{d} V^{*}$.

Let $I \in \mathcal{I}$ be any strictly increasing $d$-tuple, and define the notation $e_{I}^{*}:=e_{i_{1}}^{*} \wedge \cdots \wedge e_{i_{d}}^{*}$, so that

$$
e_{I}^{*} \stackrel{\phi}{\longmapsto} \psi_{e_{I}^{*}}\left(v_{1} \wedge \cdots \wedge v_{d}\right)=\sum_{\sigma \in S_{d}} \operatorname{sign}(\sigma) \prod_{j=1}^{d} e_{i_{j}}^{*}\left(v_{\sigma(j)}\right)= \begin{cases}1, & \text { if } v_{1} \wedge \cdots \wedge v_{d}=e_{i_{1}} \wedge \cdots \wedge e_{i_{d}}  \tag{6.9}\\ 0, & \text { otherwise }\end{cases}
$$

Notice that we mapped a basis of $\Lambda^{d} V^{*}$ into a basis of $\left(\Lambda^{d} V\right)^{*}$, since any map $\psi \in$ $\left(\Lambda^{d} V\right)^{*}$ can be obtained by a unique linear combination of maps of the collection $\left\{\psi_{e_{I}^{*}} \in\left(\Lambda^{d} V\right)^{*}: I \in \mathcal{I}\right\}$ — which, together with Eq. (6.9), makes it a basis. This shows that the morphism $\phi$ establishes an isomorphism $\Lambda^{d} V^{*} \simeq\left(\Lambda^{d} V\right)$.

Add cool propositions for the exterior algebra: the Hodge Star Operator; the isomorphisms

$$
\begin{aligned}
\operatorname{Sym}^{d}(V \oplus W) & \simeq \bigoplus_{i=0}^{d} \operatorname{Sym}^{i}(V) \otimes \operatorname{Sym}^{d-i}(W) \\
\Lambda^{d}(V \oplus W) & \simeq \bigoplus_{i=0}^{d} \Lambda^{i}(V) \otimes \Lambda^{d-i}(W)
\end{aligned}
$$

After creating sections on affine and projective geometry, introduce grassmanian varieties.

### 6.6 Determinants

Definition 6.6.1. Let $f: V \rightarrow L$ be a $k$-linear map. The map $f$ naturally induces a $k$-linear pullback $f^{\wedge d}: \Lambda^{d} V \rightarrow \Lambda^{d} L$ which is defined by

$$
v_{1} \wedge \cdots \wedge v_{d} \stackrel{f^{\wedge d}}{\longmapsto} f\left(v_{1}\right) \wedge \cdots \wedge f\left(v_{d}\right) .
$$

Proposition 6.6.2. Let $f: V \rightarrow W$ and $g: W \rightarrow L$ be $k$-linear maps. Then the composition $g f: V \rightarrow L$ satisfy

$$
(g f)^{\wedge d}=g^{\wedge d} f^{\wedge d}
$$

Proof. Notice that for any given $v=v_{1} \wedge \cdots \wedge v_{d} \in \Lambda^{d} V$ we have

$$
\begin{equation*}
(g f)^{\wedge d}(v)=(g f)\left(v_{1}\right) \wedge \cdots \wedge(g f)\left(v_{d}\right)=g\left(f\left(v_{1}\right)\right) \wedge \cdots \wedge g\left(f\left(v_{2}\right)\right)=\left(g^{\wedge d} f^{\wedge d}\right)(v) \tag{দ}
\end{equation*}
$$

Therefore $(g f)^{\wedge d}=g^{\wedge d} f^{\wedge d}$.
Definition 6.6.3 (Determinant). Let $V$ be an $n$-dimensional $k$-vector space and $f \in$ $\operatorname{End}(V)$. Since $\Lambda^{n} V$ is 1-dimensional, the induced map $f^{\wedge n} \in \operatorname{End}\left(\Lambda^{n} V\right)$ is a multiplication by some scalar in $\Lambda^{n} V$. We define the determinant of the $k$-linear endomorphism $f$ as the element $\operatorname{det} f \in \Lambda^{n} V$ such that

$$
f^{\wedge n}(\omega)=\operatorname{det}(f) \omega
$$

Proposition 6.6.4 (Composition determinant). Let $f, g \in \operatorname{End}(V)$ where $V$ is a $n$ dimensional vector space, then $\operatorname{det}(g f)=\operatorname{det} g \operatorname{det} f$.

Proof. From Proposition 6.6.2, $(g f)^{\wedge n}(\omega)=g^{\wedge n}\left(f^{\wedge n}(\omega)\right)$. From Definition 6.6.3,

$$
\operatorname{det}(g f) \omega=\operatorname{det}(g) \operatorname{det}(f) \omega
$$

thus $\operatorname{det}(g f)=\operatorname{det}(g) \operatorname{det}(f)$.
Proposition 6.6.5 (Identity determinant). Let $V$ be an $n$-dimensional vector space. The identity $\operatorname{map}^{\operatorname{id}}{ }_{V} \in \operatorname{End}(V)$ is such that $\operatorname{det}\left(\mathrm{id}_{V}\right)=1$.

Proof. Notice that $\mathrm{id}_{V}^{\wedge n}(\omega)=\omega$ and from Definition 6.6.3 we have $\mathrm{id}_{V}^{\wedge n}(\omega)=\operatorname{det}\left(\mathrm{id}_{V}\right) \omega$. Hence $\operatorname{det}\left(\mathrm{id}_{V}\right)=1$.

Proposition 6.6.6 (Dual determinant). Let $V$ be a finite dimensional vector space and $f \in \operatorname{End}(V)$. Then $\operatorname{det} f=\operatorname{det} f^{*}$, where $f^{*}$ is the dual map of $f$.

Proof.

## Dual determinant proof

Lemma 6.6.7. Let $V$ be an $n$-dimensional vector space. If $f \in \operatorname{End}(V)$ is not surjective, then $\operatorname{det} f=0$.

Proof. Let $\operatorname{im} f=W \subseteq V$. Since $f$ is not surjective, then $W$ is a subspace with $\operatorname{dim}(W)<n$, hence $\Lambda^{n} W=0$ because any collection with a number of elements greater than $\operatorname{dim}(W)$ is linearly dependent on $W$, and from Proposition 6.5.13 their wedge product is equal to zero. Given any non-zero $\omega=v_{1} \wedge \cdots \wedge v_{n} \in \Lambda^{n} V$ we find that $f^{\wedge n}(\omega)=f\left(v_{1}\right) \wedge \cdots \wedge f\left(v_{n}\right) \in \Lambda^{n} W=0$, since im $f=W$, and therefore $f^{\wedge n}(\omega)=0$. Moreover, since $f^{\wedge n}(\omega)=\operatorname{det}(f) \omega$ and $\omega \neq 0$, it follows that $\operatorname{det}(f)=0$.

Proposition 6.6.8 (Isomorphism determinant). Let $f \in \operatorname{End}(V)$ where $V$ is an $n$ dimensional vector space. Then $f$ is an isomorphism if and only if $\operatorname{det} f \neq 0$.

Proof. Let $\left\{v_{1}, \ldots, v_{n}\right\}$ be a basis for $V$. From Proposition 6.5.13, $v=v_{1} \wedge \cdots \wedge v_{n} \neq 0$. Suppose first that $f$ is an isomorphism, then $f^{\wedge n}(v)=\operatorname{det}(f) v i m p l i e s \operatorname{det} f \neq 0$.

Suppose now that $\operatorname{det} f \neq 0$. From Lemma 6.6 .7 we find that $f$ is surjective. Let $\omega \in \Lambda^{n} V$ be such that $f^{\wedge n}(\omega)=0$, then from the fact that $f^{\wedge n}(\omega)=\operatorname{det}(f) \omega$ and $\operatorname{det} f \neq 0$, it follows that $\omega=0$. Hence $\operatorname{ker} f^{\wedge n}=0$ and therefore $f^{\wedge n}$ is injective. This proves that $f^{\wedge n}$ is an isomorphism.

Proposition 6.6.9 (Matrix determinant). Let $V$ be an $n$-dimensional $k$-vector space and $f \in \operatorname{End}(V)$. Let $A: k^{n} \rightarrow k^{n}$ be the matrix representation of $f$ and $a_{i, j} \in k$ be the entries of $A$, where $1 \leqslant i, j \leqslant n$. Then

$$
\operatorname{det} A=\sum_{\sigma \in S_{n}} \operatorname{sign}(\sigma) a_{\sigma(1), 1} a_{\sigma(2), 2} \ldots a_{\sigma(n), n}
$$

Proof. Let me show you something interesting. Suppose $f: V \rightarrow L$ (the specified case for Proposition 6.6 .9 is $L=V$ ). I want to show you that in order to define the determinant we'll need $f$ to be an endomorphism in $V$, otherwise the determinant cannot be fully well-defined. Let $A=\left[a_{i, j}\right]$ with respect to the basis $\left\{v_{j}\right\}_{j=1}^{n}$ of $V$. From Definition 6.6 .3 we have that

$$
\operatorname{det}(A) v_{1} \wedge \cdots \wedge v_{n}=A v_{1} \wedge \cdots \wedge A v_{n}
$$

since $A v_{j}=\sum_{i=1}^{n} a_{i, j} v_{i}$ (see Definition 5.2.3) then we can substitute to the previous equation to obtain
$\operatorname{det}(A) v_{1} \wedge \cdots \wedge v_{n}=\sum_{i_{1}=1}^{n} a_{i_{1}, 1} v_{i_{1}} \wedge \cdots \wedge \sum_{i_{n}=1}^{n} a_{i_{n}, n} v_{i_{n}}=\sum_{1 \leqslant i_{1}, \ldots, i_{n} \leqslant n} \prod_{j=1}^{n} a_{i_{j}, j}\left(v_{i_{1}} \wedge \cdots \wedge v_{i_{n}}\right)$.
However, notice that if $i_{j}=i_{j^{\prime}}$ for some $1 \leqslant j, j^{\prime} \leqslant n$ then $v_{i_{1}} \wedge \cdots \wedge v_{i_{n}}=0$, from antisymmetry. Therefore we can write Eq. (6.10) as a sum of permutations of the set $\left\{i_{1}, \ldots, i_{n}\right\}=\{1, \ldots, n\}$, that is

$$
\begin{aligned}
\operatorname{det}(A) v_{1} \wedge \cdots \wedge v_{n} & =\sum_{\sigma \in S_{n}} a_{\sigma(1), 1} \ldots a_{\sigma(n), n}\left(v_{\sigma(1)} \wedge \cdots \wedge v_{\sigma(n)}\right) \\
& =\sum_{\sigma \in S_{n}} \operatorname{sign}(\sigma) a_{\sigma(1), 1} \ldots a_{\sigma(n), n}\left(v_{1} \wedge \cdots \wedge v_{n}\right)
\end{aligned}
$$

Since $v_{1} \wedge \cdots \wedge v_{n} \neq 0$, it follows that

$$
\operatorname{det} A=\sum_{\sigma \in S_{n}} \operatorname{sign}(\sigma) a_{\sigma(1), 1} a_{\sigma(2), 2} \ldots a_{\sigma(n), n}
$$

Corollary 6.6.10. Another equivalent way of writing the determinant of $A$ is

$$
\operatorname{det} A=\sum_{\sigma \in S_{n}} \operatorname{sign}(\sigma) a_{1, \sigma(1)} a_{2, \sigma(2)} \ldots a_{n, \sigma(n)}
$$

Proposition 6.6.11. Let $V$ be an $n$-dimensional $k$-vector space, and $\left(v_{1}, \ldots, v_{n}\right)$ be a basis for $V$, and $\omega \in \Lambda^{n} V$. Define, for every $1 \leqslant i \leqslant n$ the elements $w_{i}:=\sum_{j=1}^{n} \alpha_{i j} v_{j}$. Then

$$
\omega\left(w_{1}, \ldots, w_{n}\right)=\operatorname{det}\left[\alpha_{i j}\right] \omega\left(v_{1}, \ldots, v_{n}\right)
$$

where $\left[\alpha_{i j}\right]$ is the matrix composed of the coefficients $\alpha_{i j} \in k$ for $1 \leqslant i, j \leqslant n$.
Proof. This is simply an application of the pullback operation, notice that every $i$-th argument can be seen as the $i$-th row of the following resulting vector

$$
\left[\begin{array}{c}
\sum_{j=1}^{n} \alpha_{1 j} v_{j} \\
\vdots \\
\sum_{j=1}^{n} \alpha_{n j} v_{j}
\end{array}\right]=\left[\begin{array}{ccc}
\alpha_{11} & \ldots & \alpha_{1 n} \\
\vdots & \ddots & \vdots \\
\alpha_{n 1} & \ldots & \alpha_{n n}
\end{array}\right]\left[\begin{array}{c}
v_{1} \\
\vdots \\
v_{n}
\end{array}\right] .
$$

Thus $\omega\left(w_{1}, \ldots, w_{n}\right)=\left[\alpha_{i j}\right]^{\wedge n}\left(w\left(v_{1}, \ldots, v_{n}\right)\right)=\operatorname{det}\left[\alpha_{i j}\right] \omega\left(v_{1}, \ldots, v_{n}\right)$.
Definition 6.6.12 (Orientation). Let $V$ be an $n$-dimensional $k$-vector space, and let $\mathcal{B}$ be the collection of all basis for the vector space $V$. We define the following equivalence relation on $\mathcal{B}$ : two basis $B, B^{\prime} \in \mathcal{B}$ are equivalent said to be equivalent, $B \sim B^{\prime}$, if and only if the change of basis matrix $C$ from $B$ to $B^{\prime}$ (or vice versa) has $\operatorname{det} C>0$. Since every change of basis matrix is an isomorphism, $\operatorname{det} C$ can only be either strictly positive or strictly negative, therefore, the quotient $\mathcal{B} / \sim$ splits the basis of $V$ into two distinguished classes, both classes are said to define an orientation for $V$. If $[v]$ is an orientation class for $V$, we commonly refer to the opposite orientation as $-[v]$.

If $\phi: V \xrightarrow{\leftrightharpoons} L$ is an isomorphism of $k$-vector spaces, and $[v]$ is an orientation for $V$, then [ $\phi v$ ] is the induced orientation for $W$.

An oriented $k$-vector space is a pair $(V,[v])$, where $V$ is a $k$-vector space and $[v]$ is an orientation class for $V$. An isomorphism between oriented $k$-vector spaces $\psi:(V,[v]) \xrightarrow{\simeq}(L,[\ell])$ isa said to be orientation preserving if $[\psi v]=[\ell]$, otherwise we say that $\psi$ is order reversing.

Definition 6.6.13 (Standard euclidean orientation). The standard orientation on the euclidean space $\mathbf{R}^{n}$ is given by the basis $\left[e_{1}, \ldots, e_{n}\right]$ where $e_{j}=\left(\delta_{i j}\right)_{i=1}^{n}$ is the $j$-th unit vector of $\mathbf{R}^{n}$.

Corollary 6.6.14. Let $V$ be an $n$-dimensional $k$-vector space endowed with an inner product $\langle-,-\rangle: V \times V \rightarrow k$. Let $\omega \in \Lambda^{n} V$ be non-zero. Then, there is a unique orientation $\mu$ for $V$, for which $\mu=\left[v_{1}, \ldots, v_{n}\right]$ if and only if $\omega\left(v_{1}, \ldots, v_{n}\right)>0-$ where $\left(v_{1}, \ldots, v_{n}\right)$ is a basis for $V$.

Proof. Let $B:=\left(v_{j}\right)_{j=1}^{n}$ and $B^{\prime}:=\left(w_{j}\right)_{j=1}^{n}$ be orthonormal basis for $V$ with respect to the given inner product, and $C$ be the change of basis matrix from $B$ to $B^{\prime}$ - that is, $C$ is composed of entries $a_{i j} \in k$ such that $w_{i}=\sum_{j=1}^{n} a_{i j} v_{j}$. Since the given basis are orthonormal, we have

$$
\delta_{i j}=\left\langle w_{i}, w_{j}\right\rangle=\sum_{k, \ell=1}^{n} a_{i k} a_{j \ell}\left\langle v_{k}, v_{\ell}\right\rangle=\sum_{k=1}^{n} a_{i k} a_{j k}:=b_{i j} .
$$

Notice that $\left[b_{i j}\right]_{i, j=1}^{n}$ is simply the matrix resulting from the product of $C$ with its transpose $C^{*}$ - therefore, $C C^{*}=$ id and hence $\operatorname{det} C= \pm 1$. Regarding Proposition 6.6.11, we find that, if $\omega \in \Lambda^{n} V$ is such that $\omega\left(v_{1}, \ldots, v_{n}\right)= \pm 1$, then necessarily $\omega\left(w_{1}, \ldots, w_{n}\right)= \pm 1$ - for instance, if $\left(v_{1}, \ldots, v_{n}\right)$ is chosen so that $\omega\left(v_{1}, \ldots, v_{n}\right)=1$, for some $\omega \in \Lambda^{n} V$, then this $\omega$ is unique.

If $\mu$ is an orientation for $V$, it's clear that if $\mu=\left[v_{1}, \ldots, v_{n}\right]$ if and only if $\omega\left(v_{1}, \ldots, v_{n}\right)>$ 0 (the $\omega \in \Lambda^{n} V$ is unique by our last discussion), then $\mu$ is necessarily a unique orientation with such property.

Definition 6.6.15 (Volume element). If $V$ is a $n$-dimensional vector space endowed with an inner product $\langle-,-\rangle: V \times V \rightarrow k$. Let $\mu$ be an orientation for $V$. The form $\omega \in \Lambda^{n} V$, described in Corollary 6.6.14, is called the volume element of $V$ determined by the inner product $\langle-,-\rangle$ and orientation $\mu$.

## Useful determinant theorems, matrix determinant, etc.

Write about inverse matrix calculations

## Chapter 7

## Group Theory

### 7.1 Welcome to the Group

Lets recall Definition 1.2.16 and demystify it in the following definition.
Definition 7.1.1 (Group). A group $G$ is a groupoid $G$ with one object *. The elements of the group are the morphisms $\operatorname{Aut}_{G}(*)$. From the axioms contained in Definition 1.2.1, the graph $G$ contains the following data:
(G1) An identity element $e=\mathrm{id}_{*}$.
(G2) An associative binary operation $G \times G \rightarrow G$, commonly denoted by juxtaposition.
(G3) For all $g \in G$ there exists an inverse element $g^{-1} \in G$ such that $g g^{-1}=e=g^{-1} g$.

## Some Basic Laws

The following three propositions where already proved in a more general setting, but I'll rewrite them in this particular context for the sake of completeness.

Proposition 7.1.2. The identity element of a group is unique.
Proof. See Corollary 1.2.5.
Proposition 7.1.3. The inverse of an element of a group is unique.
Proof. See Proposition 1.2.7.
Proposition 7.1.4. Let $G$ be a group and $g, h \in G$, then $(g h)^{-1}=h^{-1} g^{-1}$.
Proof. $\left(h^{-1} g^{-1}\right)(g h)=h^{-1}\left(g^{-1} g\right) h=h^{-1} e h=h^{-1} h=e$ hence $h^{-1} g^{-1}=(g h)^{-1}$.
Proposition 7.1.5 (Cancellation). Let $G$ be a group and elements $a, b, c \in G$. If $a c=b c$ then $a=b$.

Proof. Notice that $a=a e=(a c) c^{-1}=(b c) c^{-1}=b e=b$.

Proposition 7.1.6. Let $G$ be a group. If $g \in G$, then the collection $\{g h: h \in G\}$ is equal to $G$.

Proof. Denote $G^{\prime}:=\{g h: h \in G\}$. Clearly we have $G^{\prime} \subseteq G$. On the other hand, if $\ell \in G$, the element $g\left(g^{-1} \ell\right)=e \ell=\ell \in G^{\prime}$, hence $G \subseteq G^{\prime}$. Thus $G=G^{\prime}$.

Definition 7.1.7 (Commutative group). A group $G$ is said to be commutative (or abelian) if for all $g, h \in G$ we have $g h=h g$.

Corollary 7.1.8. Let $G$ be a group such that for all $g \in G, g^{2}=e$. Then $G$ is a commutative group.

Proof. Let $g, h \in G$ be any elements, then $(g h)(h g)=g h^{2} g=g e g=g^{2}=e=(g h)(g h)$ and from cancellation law we find that $g h=h g$.

## Subgroups

Definition 7.1.9 (Subgroup). Let $G$ be a group. A subgroup $H$ of $G$ is a collection of elements $H \subseteq G$ containing the identity element, and such that composition of elements of $H$ and inverses are closed in $H$ - that is, given $x, y \in H$, then $x y \in H$ and $x^{-1}, y^{-1} \in H$. We say that a subgroup is trivial if it consists only of the identity element.

Corollary 7.1.10 (Intersection of subgroups). Let $G$ be a group and consider a nonempty collection of subgroups $\left\{G_{j}\right\}_{j \in J}$. Then, the intersection $\bigcap_{j \in J} G_{j}$ together with the group operation inherited from $G$ is a subgroup of $G$.

Proof. Let $g \in \bigcap_{j \in J} G_{j}$ be any element, then $g \in G_{j}$ for every $j \in J$, and since $G_{j}$ is a subgroup, then $g^{-1} \in G_{j}$ - that is, $g^{-1} \in \bigcap_{j \in J} G_{j}$. Let $g, h \in \bigcap_{j \in J} G_{j}$ be any two elements, then $g h \in G_{j}$ for each $j \in G$ and hence $g h \in \bigcap_{j \in J} G_{j}$.

The following notion will accompany us for the whole journey, the idea of taking elements of any set and inducing a group structure from it, using the set elements to generate this new group.

Definition 7.1.11 (Groups and generators). Let $S$ be a set and $G$ be a group. If every element of $G$ can be written as the product of finitely many powers of the elements of $S$, then we say that $G$ is generated by $S$, and the elements $s \in S$ are called generators of $G$. Moreover, if $S$ is finite, we naturally say that $G$ is finitely generated. Sometimes we denote this by $G=\langle S\rangle$. On the other hand, if $\varphi: S \rightarrow G$ is a set-function, $G$ is said to be generated by $\varphi$ if the image $\operatorname{im} \varphi$ generates the group $G$.

## Orders

Definition 7.1.12 (Order of an element). Let $G$ be a group and $g \in G$ be any element. We say that $g$ has finite order if there exists $n \in \mathbf{Z}_{>0}$ such that $g^{n}=e$. The order $|g|$ of the element $g$ is defined as the smallest such positive integer. If $g$ does not have a finite order, it is common to write $|g|=\infty$.

Lemma 7.1.13. Let $G$ be a group and $g \in G$ be an element with finite order. Then $g^{n}=e$ for some $n \in \mathbf{Z}_{>0}$ if and only if $|g|$ divides $n$.

Proof. Since $|g| \leqslant n$, define $m \in \mathbf{Z}_{>0}$ such that $n-m|g| \geqslant 0$ and $n-(m+1)|g|<0$. Define $r=n-m|g|$ to be the remainder, hence $r<|g|$. Our goal is to show that $r=0$. Notice that $g^{r}=g^{n} g^{-|g| m}=e e^{-m}=e$, which can only be the case for $r=0$, since $|g|$ is defined to be the least positive integer such that $g^{|g|}=e$. This proves that $m|g|=n$.

For the second part, suppose $n$ is a multiple of $|g|$ and denote it by $m|g|=n$. Then $g^{m|g|}=e^{m}=e$.

Definition 7.1.14 (Order of a group). Let $G$ group of finite number of elements. We define the order of $G$ to be the number of its elements and denote it by $|G|$. If $G$ is an infinite group, then $|G|=\infty$.

## Order of Products

Proposition 7.1.15 (Order of the power). Let $G$ be a group and $g \in G$ be an element with finite order. Then for all $m \in \mathbf{Z}_{\geqslant 0}$ the element $g^{m}$ has finite order. For $m=0$ we have $\left|g^{m}\right|=1$, for $m>0$ we have

$$
\left|g^{m}\right|=\frac{\operatorname{lcm}(m,|g|)}{m}=\frac{|g|}{\operatorname{gcd}(m,|g|)}
$$

Proof. From divisibility arguments, we have $\operatorname{lcm}(a, b) \operatorname{gcd}(a, b)=a b$ for integers $a$ and $b$, hence the second equality is justified. We prove the equality $\left|g^{m}\right|=\frac{\operatorname{lcm}(m,|g|)}{m}$. For the sake of notation, let $d:=\left|g^{m}\right|$. Notice that $g^{m d}=e$ and hence $|g|$ divides $m d$. Since $d$ is the least positive integer with such property, it follows that $m d$ is the least common multiple of $m$ and $|g|$. Therefore $m\left|g^{m}\right|=\operatorname{lcm}(m,|g|)$, which proves the equation.

Proposition 7.1.16. Let $G$ be a group, then for any $g, h \in G$ we have $|g h|=|h g|$.
Proof. First, let $x, y \in G$ be any elements, we prove that $\left|y x y^{-1}\right|=|x|$. Notice that for any $n \geqslant 1$ we have $\left(y x y^{-1}\right)^{n}=y x^{n} y^{-1}$, hence the least element that annihilates the product $y x y^{-1}$ is the order of $x$ - that is, $\left|y x y^{-1}\right|=x$. In particular, $h g=g^{-1} g h g$, hence $|h g|=\left|g^{-1}(g h) g\right|=|g h|$.

Proposition 7.1.17. Let $G$ be a group and $g, h \in G$ be such that $g h=h g$. Then $|g h|$ divides $\operatorname{lcm}(|g|,|h|)$.

Proof. Let $n$ be a common multiple of $|g|$ and $|h|$, then $g^{n}=h^{n}=e$ from Lemma 7.1.13. Notice that the commutative property $g h=h g$ allow us to permute the terms of $g^{n} h^{n}=e$ in order to obtain $g^{n} h^{n}=(g h)^{n}=e$. In particular, since lcm $(|g|,|h|)$ is a common multiple of $|g|$ and $|h|$, we find that $(g h)^{\operatorname{lcm}(|g|,|h|)}=e$ and again from Lemma 7.1.13 we find that $|g h|$ divides $\operatorname{lcm}(|g|,|h|)$

Lemma 7.1.18. Let $G$ be a group and $g, h \in G$ commute $-g h=h g$. If $\operatorname{gcd}(|g|,|h|)=1$, then $|g h|=|g||h|$.

Proof. Let $|g h|=\ell,|g|=m$ and $|h|=n$. From Proposition 7.1.17, we have that $\ell \mid \operatorname{lcm}(m, n)$, and since $m n=\operatorname{lcm}(m, n) \operatorname{gcd}(m, n)=\operatorname{lcm}(m, n)$, then $\ell \mid m n$, which implies that $\ell \leqslant m n$. Moreover, since the elements commute, $(g h)^{\ell}=g^{\ell} h^{\ell}=e$ hence $g^{\ell}=\left(h^{\ell}\right)^{-1}$. From Proposition 7.1.15, $\left|g^{\ell}\right|=\frac{|g|}{\operatorname{gcd}(\ell,|g|)}=\frac{m}{m}=1$ and equivalently for $h$ we have $\left|h^{\ell}\right|=1$. This shows that $g^{\ell}=h^{\ell}=e$ and therefore both $m$ and $n$ divide $\ell$, hence so does the product $m n$, thus $m n \leqslant \ell$. This completes the proof that $\ell=m n$.

Definition 7.1.19 (Maximal finite order). Let $G$ be a group. An element $g \in G$ is said to be of maximal finite order if its order is finite and for all $h \in G$ with finite order, we have $|h| \leqslant|g|$.

Proposition 7.1.20. Let $G$ be a commutative group and $g \in G$ be of maximal finite order. If $h \in G$ has finite order, then $|h|$ divides $|g|$.

Proof. Define the notation $|g|=m$ and $|h|=n$. Let $P=\left(p_{j}\right)_{j}$ be a finite sequence containing all primes such that less than or equal to $m$ and define a finite sequence of integers of same length $A=\left(a_{j}\right)_{j}$ such that $m=\prod_{j} p_{j}^{a_{j}}$. Since $n \leqslant m$ it follows that there also exists a finite sequence of integers $B=\left(b_{j}\right)_{j}$ such that $n=\prod_{j} p_{j}^{b_{j}}$. Suppose for the sake of contradiction that $n$ doesn't divide $m$, so that there exists an index $k$ such that $a_{k}<b_{k}$ - from the fact that $\frac{m}{n}=\prod_{j} p_{j}^{a_{j}-b_{j}}$. Consider now the order - following from Lemma 7.1.18:

$$
\left|g^{\left(p_{k}^{a_{k}}\right)} h^{\left(n / p_{k}^{b_{k}}\right)}\right|=\left|g^{\left(p_{k}^{a_{k}}\right)}\right|\left|h^{\left(n / p_{k}^{b_{k}}\right)}\right|=\frac{m}{\operatorname{gcd}\left(m, p_{k}^{a_{k}}\right)} \frac{n}{\operatorname{gcd}\left(n, n / p_{k}^{b_{k}}\right)}=\frac{m}{p_{k}^{a_{k}}} \frac{n}{n / p_{k}^{b_{k}}}=m p_{k}^{b_{k}-a_{k}}
$$

and since $b_{k}>a_{k}$, it follows that $m p_{k}^{b_{k}-a_{k}}>m$ and hence there is a contradiction since we assumed that $m$ was the the maximal finite order of the group. We conclude that there does not exist $k$ for which $b_{k}$ is less than $a_{k}$ - thus $n$ divides $m$.

## Finite Groups and Elements of Order 2

Lemma 7.1.21 (Order 2 elements implies commutative). Let $G$ be a group such that all non-identity elements have order 2 . Then $G$ is commutative.

Proof. Let $g, h \in G$, notice that $|g h|=2$ from Proposition 7.1.17 hence $g h \in G$, then $(g h)^{2}=e$ and therefore $g h=g^{-1} h^{-1}=g^{-1}(g h g h) h^{-1}=h g$ commutes.

Lemma 7.1.22. Let $G$ be a finite group such that any element has order at most 2. Let $H$ be any subgroup of $G$, for any $t \in G \backslash H$ consider the collection $T=H \cup\{h t: h \in H\}$, then $T$ is a subgroup of $G$ with $|T|=2|H|$.

Proof. Since $t \notin H$, any element $h t \in T$ with $h \in H$ is such that $h t \notin H$. The number of distinct elements of the form is $|H|$ - one for each $h \in H$ - hence $|T|=|H|+|H|=$ $2|H|$.

Lemma 7.1.23 (Order $2^{n}$ ). Let $G$ be a finite group such that any element has order at most 2. The order of $G$ is of the form $2^{n}$ for some $n \geqslant 0$. Moreover, if $|G|>1$, then there exists a subgroup $H$ of $G$ with order $|H|=2^{n-1}$.

Proof. We create a recursive algorithm to find the collection of elements of $G$ that generate any other element contained in $G$ :

1. (Base case) If $G=H_{j}$, return $H_{j}$.
2. (Recursion) Let $g \in G \backslash H_{j}$ and construct $H_{j+1}=H_{j} \cup\left\{h g: h \in H_{j}\right\}$, recursively call the algorithm with $H_{j+1}$.

Such algorithm is certain to terminate since $|G|$ is finite. Notice that the second step always doubles the order of the set $H_{j}$, so that - if $H_{n}$ is the result of the algorithm - then $\left|H_{n}\right|=2^{n}$. This shows that $|G|=\left|H_{n}\right|=2^{n}$. The second part of the statement follows immediately from the construction of the algorithm.

Proposition 7.1.24. Let $G$ be a commutative group, if there exists exactly one element of order 2 - say $f \in G$ - then the product of all elements of the group is

$$
\prod_{g \in G} g=f
$$

Otherwise, we have $\prod_{g \in G} g=e$.
Proof. Since $G$ is finite, let $G=\left\{e, g_{1}, g_{1}^{-1}, \ldots, g_{n}, g_{n}^{-1}\right\}$. Suppose there exists one and only one element of order 2 and denote it by $f$ - that is, $f=f^{-1}$. Since the group is commutative, we can rearrange the product of elements so that we can take the pairwise product of each element and its inverse - this being possible only in the case where $g \neq g^{-1}$ - taking the product we find

$$
\prod_{g \in G} g=e\left(g_{1} g_{1}^{-1}\right) \ldots\left(g_{j} g_{j}^{-1}\right) f \ldots\left(g_{j+2} g_{j+2}^{-1}\right)\left(g_{n} g_{n}^{-1}\right)=e^{j+1} f e^{n-(j+1)}=f
$$

since $f$ has no inverse in $G \backslash\{f\}$.
Let there be no element with order 2 in $G$, then the rearrangement of pairwise element and respective inverse is possible with each $g \in G$, making

$$
\prod_{g \in G} g=e\left(g_{1} g_{1}^{-1}\right) \ldots\left(g_{j} g_{j}^{-1}\right) \ldots\left(g_{n} g_{n}^{-1}\right)=e^{n+1}=e
$$

Suppose there exists $m>1$ distinct elements in $G$ with order 2 and define $T=\{f \in$ $G:|f| \leqslant 2\}$ and $S=G \backslash T$. From the last case we know that $\prod_{s \in S} s=e$. Note that $T$ forms a subgroup of $G$ :

- $e \in T$.
- If $g, h \in T$, then from Proposition 7.1.17 $|g h|$ divides $1 \mathrm{~cm}(|g|,|h|) \leqslant 2$ hence $|g h| \leqslant 2$ and $g h \in T$.
- Since $g \in T$ implies $g g=e$ then $g=g^{-1}$ and $\left|g^{-1}\right|=2$, so $g^{-1} \in T$.

Moreover, the order of the group $T$ has to be of the form $2^{k}$ for some $k \geqslant 2$, from Lemma 7.1.23. From the same lemma, choose a subgroup $H$ with order $2^{k-1}$ and take $u \in T \backslash H$, so that $T=H \cup\{h u: h \in H\}$ - such $u$ exists from the algorithm constructed in the lemma. Since $T$ is commutative (see Lemma 7.1.21), we can write

$$
\prod_{t \in T} t=\prod_{h \in H} h \prod_{h \in H} h u=\prod_{h \in H} u h^{2}=\prod_{h \in H} u=u^{2^{k-1}}=\left(u^{2}\right)^{2^{k-2}}=e .
$$

Therefore we can finally conclude that

$$
\begin{equation*}
\prod_{g \in G} g=\prod_{s \in S} s \prod_{t \in T} t=e \tag{দ}
\end{equation*}
$$

### 7.2 Examples of Groups

## Symmetry Group

Definition 7.2.1 (Symmetric groups). Let $A \in$ Set. The symmetric group of $A-$ also referred to as the permutation group of $A-\operatorname{Aut}_{\text {set }}(A)$, is denoted by $S_{A}$. The symmetric group of the range set $\{1, \ldots, n\}$ is denoted $S_{n}$.

Notation 7.2.2 (Permutations). A permutation $\sigma \in S_{n}$ is denoted by a table of the form

$$
\sigma=\left(\begin{array}{ccccc}
1 & 2 & \ldots & n-1 & n \\
\sigma(1) & \sigma(2) & \ldots & \sigma(n-1) & \sigma(n)
\end{array}\right)
$$

Remark 7.2.3 (Convention). If $\sigma, \tau \in S_{n}$, we write the composition $\sigma \tau$ to denote the permutation

$$
i \stackrel{\sigma \tau}{\longmapsto} \sigma(\tau(i)) .
$$

That is, it follows the same order of composition for maps.
Proposition 7.2.4. There exists an embedding $f: S_{n} \hookrightarrow M_{n \times n}(\{0,1\})$ — where $M_{n \times n}(\{0,1\})$ is the collection of $n \times n$ matrices with entries assuming the value of 1 or 0 . Moreover, if $f(\sigma)=M_{\sigma}$ and $f(\tau)=M_{\tau}$ then $f(\sigma \tau)=M_{\sigma} M_{\tau}$.

Proof. Let $\sigma \in S_{n}$ be any permutation. We define $f$ as the mapping

$$
\sigma \stackrel{f}{\longmapsto} M_{\sigma}=\left[\begin{array}{ccc}
\delta_{1 \sigma(1)} & \ldots & \delta_{1 \sigma(n)} \\
\vdots & \ddots & \vdots \\
\delta_{n \sigma(1)} & \ldots & \delta_{n \sigma(n)}
\end{array}\right]
$$

It should be noted that since $\sigma$ is injective, $\delta_{i \sigma(j)}$ assumes the value 1 only once in the $j$-th column. If $\sigma=\tau$ then clearly $\delta_{i \sigma(j)}=\delta_{i \tau(j)}$ for every $1 \leqslant i, j \leqslant n$, thus $f(\sigma)=f(\tau)$ - $f$ is injective.

Consider now the composition of permutations $\sigma \tau \in S_{n}$. Suppose we want to check where the $j$-th element goes to from the action of $\sigma \tau$. Notice that for all possible new positions $1 \leqslant i \leqslant n$ we have $\delta_{i \sigma \tau(j)}=\sum_{k=1}^{n} \delta_{i \sigma(k)} \delta_{k \tau(j)}$ - that is, $j$ belongs to $i$ after $\sigma \tau$ if and only if $\tau(j)=k$ and $\sigma(k)=i$ for some $1 \leqslant k \leqslant n$. We have
$f(\sigma \tau)=M_{\sigma \tau}=\left[\begin{array}{ccc}\delta_{1 \sigma \tau(1)} & \ldots & \delta_{1 \sigma \tau(n)} \\ \vdots & \ddots & \vdots \\ \delta_{n \sigma \tau(1)} & \ldots & \delta_{n \sigma \tau(n)}\end{array}\right]=\left[\begin{array}{ccc}\delta_{1 \sigma(1)} & \ldots & \delta_{1 \sigma(n)} \\ \vdots & \ddots & \vdots \\ \delta_{n \sigma(1)} & \ldots & \delta_{n \sigma(n)}\end{array}\right]\left[\begin{array}{ccc}\delta_{1 \tau(1)} & \ldots & \delta_{1 \tau(n)} \\ \vdots & \ddots & \vdots \\ \delta_{n \tau(1)} & \ldots & \delta_{n \tau(n)}\end{array}\right]=M_{\sigma} M_{\tau}$.

Proposition 7.2.5. The symmetry group $S_{n}$ contains elements of all orders $d$ for $1 \leqslant$ $d \leqslant n$.

Proof. Let $d$ be any integer in the range $1 \leqslant d \leqslant n$. Consider the subgroup $S_{d} \subseteq$ Aut $_{\text {set }}(\mathbf{Z} / n \mathbf{Z})$, composed of permutations $\sigma \in S_{d}$ such that $\sigma\left([i]_{n}\right)=[i]_{n}$ for all $i>d$. Notice that the permutation $\tau$ given by $\tau\left([i]_{n}\right)=[i+1]_{n}$ for all $i \leqslant d$ and $\tau\left([j]_{n}\right)=[j]_{n}$ is an element of $S_{d}$ and $|\tau|=d$. Thus $\tau$ is an element of order $d$ in $S_{n}$.

Corollary 7.2.6. For every $n \in \mathbf{N}$, there exists an element $x \in S_{\mathbf{N}}$ with order $|x|=n$.
Proof. Let any $n \in \mathbf{N}$ and choose $m \in \mathbf{N}$ with $m \geqslant n$. From Proposition 7.2.5 we find that the group $S_{m}$ contains an element $\sigma \in S_{m}$ with order $|\sigma|=n$. We can now construct an element $\tau \in S_{\mathbf{N}}$ - defined by

$$
\tau(a)= \begin{cases}\sigma(a), & a \leqslant m \\ a, & a>m\end{cases}
$$

So clearly $|\tau|=n$.

## Permutations, Transpositions and Sign

Definition 7.2.7 (Transposition). We define a transposition on a collection $\{1, \ldots, n\}$ to be a map $\tau \in S_{n}$ such that exists indices $1 \leqslant i<j \leqslant n$ for which $\tau(i)=j$ and $\tau(j)=i$, and $\tau(k)=k$ for all $k \neq i, j$.
Proposition 7.2.8. Every permutation can be written as a composition of finitely many transpositions.

Proof. We proceed via induction on the number of points of $\{1, \ldots, n\}$. For the base case $n=2$ the composition is trivial. Assume as the induction hypothesis that for $n-1>2$ the statement is true. Now, consider $\sigma \in S_{n}$ and $i \in\{1, \ldots, n\}$ be any element in the ordered collection $I_{n}:=\{1, \ldots, n\}$. Denote $\tau_{i, j}$ the transposition that changes $i$ with $j$ and maintains unchanged the remainder of the points. Assume that $\sigma(i)=j$, then clearly $\tau_{i, j} \sigma(i)=\tau_{i, j}(j)=i$. Notice that since $\tau_{i, j} \sigma$ maintains $i$ unchanged, we can see it as a permutation of $n-1$ points (by simply ignoring the point $i$ ), hence it can be written as a composition of finitely many transpositions by the induction hypothesis. Notice that since $\sigma=\tau_{i, j}\left(\tau_{i, j} \sigma\right)$ we find that $\sigma$ can also be written as a composition of finitely many transpositions.

Definition 7.2.9 (Elementary transpositions). Let $\tau \in S_{n}$ be a transposition. We say that $\tau$ is an elementary transposition if exists $i \in\{1, \ldots, n\}$ for which $\tau(i)=i+1$ and $\tau(i+1)=i$, and for all $j \neq i$ we have $\tau(j)=j$.

Proposition 7.2.10. Every transposition can be written as a composition of finitely many elementary transpositions.

Proof. Let $\sigma \in S_{n}$ be any transposition. Denote by $\tau_{k}$ the elementary transposition $\tau_{k}(k)=k+1$ and $\tau(k+1) k$. Let $x$ be the transposed element of $\sigma$, and $\sigma(x)=y$. Without loss of generality we can assume that $y>x$. Now we write $\sigma$ as the composition beware of Remark 7.2.3

$$
\sigma=\left(\tau_{x} \tau_{x+1} \cdots \tau_{y-3} \tau_{y-2}\right)\left(\tau_{y-1} \tau_{y-1} \cdots \tau_{x+1} \tau_{x}\right)=\left(\prod_{k=x}^{y-2} \tau_{k}\right)\left(\prod_{k=0}^{(y-1)-x} \tau_{(y-1)-k}\right)
$$

Which follows from the fact that

$$
\begin{gathered}
\left(\prod_{k=0}^{(y-1)-x} \tau_{(y-1)-k}\right)(i)= \begin{cases}y, & i=x \\
i-1, & x<i \leqslant y \\
i, & i<x \text { or } y<i\end{cases} \\
\left(\prod_{k=x}^{y-2} \tau_{k}\right)(i)= \begin{cases}x, & i=y \\
i+1, & x \leqslant i<y \\
i, & i<x \text { or } y<i\end{cases}
\end{gathered}
$$

Hence the composition of both gives

$$
\left(\prod_{k=x}^{y-2} \tau_{k}\right)\left(\prod_{k=0}^{(y-1)-x} \tau_{(y-1)-k}\right)(i)= \begin{cases}y, & i=x \\ x, & i=y \\ i, & i \notin\{x, y\}\end{cases}
$$

which is equivalent to the transposition $\sigma$.
Corollary 7.2.11. Every permutation can be written as a composition of finitely many elementary transpositions.

Definition 7.2.12 (Sign). Let $\sigma \in S_{n}$ be a permutation on $n$ objects. We define the sign of $\sigma$ as a map sign: $S_{n} \rightarrow\{1,-1\}$ such that

$$
\operatorname{sign}(\sigma)=(-1)^{m}, \text { where } m:=|\{(i, j): 1 \leqslant i<j \leqslant n, \sigma(i)>\sigma(j)\}|
$$

Equivalently, if $x_{1}, \ldots, x_{n} \in k$ then we can define the sign of $\sigma$ as

$$
\operatorname{sign}(\sigma)=\prod_{i<j} \frac{x_{\sigma(i)}-x_{\sigma(j)}}{x_{i}-x_{j}}
$$

Corollary 7.2.13. If $\sigma$ can be written as an odd number of transpositions, then $\operatorname{sign}(\sigma)=$ -1 . Otherwise, if it is an even number of transpositions, we have $\operatorname{sign}(\sigma)=1$.

Proposition 7.2.14. Let $\sigma, \tau \in S_{n}$, then

$$
\operatorname{sign}(\sigma \tau)=\operatorname{sign}(\sigma) \operatorname{sign}(\tau)
$$

This implies that sign: $S_{n} \rightarrow\{1,-1\}$ is a group homomorphism.
Proof. Consider the composition $\sigma \tau$, then we have

$$
\begin{equation*}
\operatorname{sign}(\sigma \tau)=\prod_{i<j} \frac{x_{\sigma \tau(i)}-x_{\sigma \tau(j)}}{x_{i}-x_{j}} \tag{7.1}
\end{equation*}
$$

Define the following: when $\tau(i)<\tau(j)$ then $\tau(i):=p$ and $\tau(j):=q$; on the other hand, when $\tau(j)<\tau(i)$ let $\tau(i):=q$ and $\tau(j):=p$. This way we find that

$$
\begin{equation*}
\frac{x_{\sigma \tau(i)}-x_{\sigma \tau(j)}}{x_{\tau(i)}-x_{\tau(j)}}=\frac{x_{\sigma(p)}-x_{\sigma(q)}}{x_{p}-x_{q}} . \tag{7.2}
\end{equation*}
$$

From construction of the above, Eq. (7.2) yields

$$
\begin{equation*}
\operatorname{sign}(\sigma)=\prod_{p<q} \frac{x_{\sigma(p)}-x_{\sigma(q)}}{x_{p}-x_{q}} \tag{7.3}
\end{equation*}
$$

Notice that we can write Eq. (7.1) (using Eq. (7.3)) as the product

$$
\operatorname{sign}(\sigma \tau)=\prod_{i<j} \frac{x_{\sigma \tau(i)}-x_{\sigma \tau(j)}}{x_{\tau(i)}-x_{\tau(j)}} \frac{x_{\tau(i)}-x_{\tau(j)}}{x_{i}-x_{j}}=\operatorname{sign}(\sigma) \operatorname{sign}(\tau)
$$

as wanted.

## Dihedral Group

Definition 7.2.15 (Dihedral group). A dihedral group is defined as the group of isometric symmetries of regular polygons - which are rotations and reflections about a line. Given a $n$-sided regular polygon, its group of symmetries have $2 n$ elements - $n$ rotations and $n$ reflections - and we denote it by $D_{2 n}$.

Proposition 7.2.16. There exists an embedding $D_{2 n} \hookrightarrow S_{n}$.
Proof. Label the vertices of the $n$-gon by $V=\left\{[1]_{n}, \ldots,[n]_{n}\right\}=\mathbf{Z} / n \mathbf{Z}$. Notice that any element $x \in D_{2 n}$ can be described as $x \in \operatorname{Aut}_{\text {Set }}(V)$ - where the automorphism is restricted to the adjacency of the vertices, that is, if $x\left([i]_{n}\right)=[k]_{n}$, then $x([i-$ $\left.1]_{n}\right), x\left([i+1]_{n}\right) \in\left\{[k-1]_{n},[k+1]_{n}\right\}$. This shows the existence of the embedding $D_{2 n} \hookrightarrow \operatorname{Aut}_{\text {set }}(V)=S_{V}=S_{n}$.

Proposition 7.2.17. Any symmetry $x \in D_{2 n}$ can be written as $y^{a} z^{b}$ - where we choose any $y, z \in D_{2 n}$ which are, respectively, a rotation and a reflection about a line - with $0 \leqslant a<2$ and $0 \leqslant b<n$.

Proof. Notice that if $y$ is any rotation, then $|y|=n$ and if $z$ is any reflection, then $|z|=2$.
We first show that $y$ and $z$ are independent. We'll work with the injection $D_{2 n} \mapsto$ $\operatorname{Aut}_{\text {set }}(V)$, where we have the collection of vertices $V=\mathbf{Z} / n \mathbf{Z}$. Let $j \in V$ be any vertex and suppose $z(j)=k$, then by the adjacency of the vertices are maintained, implying in $z(j-1)=k-1$ and $z(j+1)=k+1$. On the other hand, if $y(j)=k$, then the adjacency of the vertices is inverted, that is, $y(j-1)=k+1$ and $y(j+1)=k-1$. Hence clearly reflections and rotations cannot be dependent.

The only possible symmetries of $D_{2 n}$ involve the maintenance or the inversion of the adjacency of each vertex - we cannot break nor deform the edges that connect each of the vertices - thus if $x \in D_{2 n}$, then $x=\prod_{(\alpha, \beta) \in I} z^{\alpha} y^{\beta}$ for some finite set $I$ with $0 \leqslant \alpha<2$ and $0 \leqslant \beta<n$. We now show that such product can be reduced. Let $j \in V$ be any vertex and assume

$$
z(j)=k \quad \text { and } \quad y(k)=\ell
$$

We now analyse the symmetries $y z$ and $z y^{m}-$ for $0 \leqslant m<n$.

- The adjacent vertices of $j$ - when subjected to the symmetry $y z$ - are obtained as

$$
y z(j-1)=y(k+1)=\ell+1 \quad \text { and } \quad y z(j+1)=y(k-1)=\ell-1
$$

- Since $y(k)=\ell, y$ is the action rotating the vertex $k$ an amount of $\ell-k$ times - where we'll adopt the convention that if $\ell-k<0$, the rotation is counterclockwise, if $\ell+k>0$ the rotation is clockwise. Therefore

$$
y(j)=j+(\ell-k)
$$

Notice that this implies in $y^{m}(j)=j+m(\ell-k)$ - we now define $t=y^{m}(j)$. Since $z(j)=k$, then $z(t)=k-m(\ell-k)$. Therefore, adjacent vertices of $j-$ when subjected to the symmetry $z y^{m}$ - are given by

$$
\begin{array}{ll}
\text { Vertex } j-1: & z y^{m}(j-1)=z(j+m(\ell-k)-1)=k-m(\ell-k)+1 \\
\text { Vertex } j+1: & z y^{m}(j+1)=z(j+m(\ell-k)+1)=k-m(\ell-k)-1
\end{array}
$$

In particular, for the case $m=n-1$ - beware of the modularity of the vertices, they lie on $\mathbf{Z} / n \mathbf{Z}$ - we get

$$
\begin{aligned}
& \text { Vertex } j-1: z y^{n-1}(j-1)=k-(n-1)(\ell-k)+1=k n-n \ell+\ell+1=\ell+1 \\
& \text { Vertex } j: \quad z y^{n-1}(j)=k-(n-1)(\ell-k)=k n-n \ell+\ell=\ell \\
& \text { Vertex } j+1: z y^{n-1}(j+1)=k-(n-1)(\ell-k)-1=k n-n \ell+\ell-1=\ell+1
\end{aligned}
$$

From this analysis we conclude the general relation $y z=z y^{n-1}$.
Lets analyse the product $\left(z^{a} y^{b}\right)\left(z^{c} y^{d}\right)$ for the different cases of $0 \leqslant a, c<2$ and $0 \leqslant b, d<n$.

- If $a=1$ and $c=0$ then $\left(z^{a} y^{b}\right)\left(z^{c} y^{d}\right)=z y^{b+d}$.
- If $a, c=0$ then

$$
\left(z^{a} y^{b}\right)\left(z^{c} y^{d}\right)=y^{b+d} .
$$

- If $a=0$ and $c=1$ then $\left(z^{a} y^{b}\right)\left(z^{c} y^{d}\right)=y^{b} z^{c} y^{d}=y^{b-1} z y^{d+(n-1)}$, by recurrence we find find finally that

$$
\left(z^{a} y^{b}\right)\left(z^{c} y^{d}\right)=z y^{d+b(n-1)} .
$$

- If $a, c=1$ then $\left(z^{a} y^{b}\right)\left(z^{c} y^{d}\right)=z y^{b} z y^{d}=z y^{b-1} z y^{d+(n-1)}$ and, by recurrence

$$
\left(z^{a} y^{b}\right)\left(z^{c} y^{d}\right)=z^{2} y^{d+b(n-1)}=y^{d+b(n-1)}
$$

This shows that the finite product $x=\prod_{(\alpha, \beta) \in I} z^{\alpha} y^{\beta}$ can be reduced - for some $0 \leqslant a<2$ and $0 \leqslant b<n$ - to $x=z^{a} y^{b}$.

Corollary 7.2.18. Let $y, z \in D_{2 n}$ be, respectively, a rotation and a reflection. The following relation are satisfied:

$$
|y|=n,|z|=2 \text {, and } y z=z y^{n-1} .
$$

The set $\{y, z\}$ generates any element of $D_{2 n}$.

## Cyclic Group

We'll initially define a cyclic group as a group generated by one element $x$ with a relation $x^{n}=e$ for some $n \in \mathbf{N}$. We'll denote such groups by $C_{n}$ - for a formal definition, see Definition 7.3.20.

Example 7.2.19. An example of a cyclic group is $(\mathbf{Z} / n \mathbf{Z},+)$, where $[1]_{n} \in \mathbf{Z} / n \mathbf{Z}$ is the element generating any elements of $\mathbf{Z} / n \mathbf{Z}$.

Proposition 7.2.20. Given an element $y \in C_{n}$ - where $x$ is the generating element of $C_{n}$ - assume $y=x^{m}$. We have that

$$
|y|=\frac{n}{\operatorname{gcd}(m, n)} .
$$

Proof. Notice that $|y|=\left|x^{m}\right|=\frac{|x|}{\operatorname{gcd}(m, x \mid)}=\frac{n}{\operatorname{gcd}(m, n)}$.
Corollary 7.2.21. The element $x \in C_{n}$ is the generator of the cyclic group if and only if

$$
\operatorname{gcd}(|x|, n)=1
$$

### 7.3 Grp Category

Definition 7.3.1 (Group morphism). Let $\left(G,{ }_{G}\right)$ and $\left(H,{ }_{H}\right)$ be groups together with their binary operation. A group morphism - also called homomorphism - is a map $\varphi:\left(G,{ }_{G}\right) \rightarrow\left(H, \cdot_{H}\right)$ such that the that the following diagram commutes


Where $\varphi \times \varphi$ is uniquely defined in Set by ?? - mapping $(g, \ell) \stackrel{\varphi \times \varphi}{\longmapsto}(\varphi(g), \varphi(\ell))$. The commutativity of such diagram can be viewed as the requirement that $\varphi$ preserves the structure coming from the binary operations - that is, for any $g, \ell \in G$

$$
\varphi\left(g \cdot{ }_{G} \ell\right)=\varphi(g) \cdot H \varphi(\ell)
$$

Definition 7.3.2 (Category of groups). The category of groups Grp consists of the collection of objects - called groups - and group morphisms between them.

Proposition 7.3.3. Grp is a category.
Proof. Let $\left(G, \cdot \cdot_{G}\right),\left(H, \cdot_{H}\right)$ and $\left(K, \cdot_{K}\right)$ be any groups. The identity $\mathrm{id}_{G}: G \rightarrow G$ is a group morphism since $\operatorname{id}_{G}(g \cdot G \ell)=g \cdot{ }_{G} \ell$ for any $g, \ell \in G$. Moreover, we can define a map

$$
f: \operatorname{Mor}_{G r p}(G, H) \times \operatorname{Mor}_{G r p}(H, K) \rightarrow \operatorname{Mor}_{G r p}(G, K)
$$

with the mapping $(\psi, \varphi) \stackrel{f}{\mapsto} \psi \varphi$ — since the following diagram commutes


In other words, for any $g, \ell \in G$ we have

$$
\psi \varphi\left(g \cdot{ }_{G} \ell\right)=\psi\left(\varphi\left(g \cdot{ }_{G} \ell\right)\right)=\psi\left(\varphi(g) \cdot{ }_{H} \varphi(\ell)\right)=\psi(\varphi(g)) \cdot K \psi(\varphi(\ell))=\psi \varphi(g) \cdot{ }_{K} \psi \varphi(\ell)
$$

Therefore $\psi \varphi \in \operatorname{Mor}_{G r p}(G, K)$. For the other part of the diagram, we have

$$
(\psi \varphi) \times(\psi \varphi)(g, \ell)=(\psi \varphi(g), \psi \varphi(\ell))=\psi \times \psi(\varphi(g), \varphi(\ell))=(\psi \times \psi)(\varphi \times \varphi)(g, \ell)
$$

Since group morphisms are maps in Set, we have that associativity is inherited. $\square$

Proposition 7.3.4. There exists a covariant forgetful functor $F:$ Grp $\rightarrow$ Set.
Proof. For objects, define $F$ as $F(G, \cdot G)=G$ - where we denoted $G$ together with its binary operation only to express that the multiplicative structure is lost in the process. Let $\varphi:\left(G, \cdot{ }_{G}\right) \rightarrow\left(H, \cdot{ }_{H}\right)$ be a group morphism, denote by $\bar{\varphi} \in \operatorname{Mor}($ Set $)$ the function $\bar{\varphi}: G \rightarrow H$ such that $\bar{\varphi}(g)=\varphi(g)$ for all $g \in G$. For such morphisms we define $F$ as $F \varphi=\bar{\varphi}: F\left(G, \cdot{ }_{G}\right) \rightarrow F\left(H, \cdot{ }_{H}\right)$.

Let $\psi \in \operatorname{Mor}_{\mathrm{Grp}}(H K)$, then we have $\overline{\psi \varphi}=\bar{\psi} \bar{\varphi}: H \rightarrow K$. Thus

$$
F(\psi \varphi)=\overline{\psi \varphi}=\bar{\psi} \bar{\varphi}=F \psi F \varphi
$$

This shows that $F$ is a covariant forgetful functor.
Proposition 7.3.5. The trivial group $* \in \operatorname{Grp}$ is the initial and final object of Grp. That is, for any $G \in$ Grp the diagram

$$
* \stackrel{\psi}{\underset{\varphi}{-}} \underset{\rightarrow}{ } G
$$

commutes for uniquely defined group morphisms $\varphi$ and $\psi$.
Proof. Let $G \in \operatorname{Grp}$ be any group. We define maps $\varphi: * \rightarrow G$ mapping $e \stackrel{\varphi}{\mapsto} e_{G}$, where $e$ is the only element of $*$ - being unique possible map $* \rightarrow G$ that preserves the group structure. Clearly, $\varphi$ is a group morphism since $\varphi(e e)=\varphi(e)=e_{G}=\varphi(e) \varphi(e)$ - this shows that $*$ is the initial object of Grp. Let $\psi: G \rightarrow *$ be a map defined by $g \stackrel{\psi}{\mapsto} e-$ which is clearly unique. Then $\psi$ is a morphism of groups, because $\psi(g h)=e=\psi(g) \psi(h)$ showing that * is the final object of Grp.

## Properties of Morphisms

Proposition 7.3.6 (Commuting on inverses). Let $\left(G, \cdot{ }_{G}\right),\left(H, \cdot{ }_{H}\right)$ be groups and consider $\varphi \in \operatorname{Mor}_{G r p}(G, H)$. Define $\operatorname{inv}_{G}: G \xrightarrow{\leftrightharpoons} G$ and $\operatorname{inv}_{H}: H \xrightarrow{\leftrightharpoons} H$ as the maps $g \stackrel{\text { inv }}{\longmapsto} g^{-1}$ and $h \stackrel{\operatorname{inv}_{H}}{\longrightarrow} h^{-1}$. Then the following diagram commutes


That is, $\varphi\left(g^{-1}\right)=\varphi(g)^{-1}$ for every $g \in G$.
Proof. Let $g \in G$ be any element, then

$$
\varphi\left(g^{-1}\right)=\varphi\left(g^{-1} e_{G}\right)=\varphi\left(g^{-1} \cdot{ }_{G} g \cdot G g^{-1}\right)=\varphi\left(g^{-1}\right) \cdot H \varphi(g) \cdot{ }_{H} \varphi\left(g^{-1}\right)
$$

applying cancellation law on the equation above we find

$$
e_{H}=\varphi(g) \cdot H \varphi\left(g^{-1}\right)
$$

Hence $\varphi\left(g^{-1}\right)=\varphi(g)^{-1}$. Moreover, this implies that $e_{G} \stackrel{\varphi}{\mapsto} e_{H}$ since

$$
\varphi\left(e_{G}\right)=\varphi\left(g \cdot{ }_{G} g^{-1}\right)=\varphi(g) \cdot H \varphi\left(g^{-1}\right)=\varphi(g) \cdot H \varphi(g)^{-1}=e_{H}
$$

Proposition 7.3.7 (Generators and unique extension). Let $G$ be a group and $S$ be a generator set for $G$. Let $H$ be any group and $f: S \rightarrow H$ be a set-function. If there exists a morphism $\phi: G \rightarrow H$ such that $\left.\phi\right|_{S}=f$, then $\phi$ is unique.

Proof. Let $\psi: G \rightarrow H$ be another morphism satisfying the condition specified above then clearly $\left.\psi\right|_{S}=f=\left.\phi\right|_{S}$, that is, $\phi$ and $\psi$ agree on $S$. Since $S$ generates $G$, every element $g \in G$ can be written as a finite product $g=\prod_{j} s_{j} \in\langle S\rangle$ thus

$$
\begin{equation*}
\psi(g)=\psi\left(\prod_{j} s_{j}\right)=\prod_{j} \psi\left(s_{j}\right)=\prod_{j} \phi\left(s_{j}\right)=\phi\left(\prod_{j} s_{j}\right)=\phi(g) \tag{4}
\end{equation*}
$$

which implies in $\psi=\phi$.
Proposition 7.3.8 (Image subgroup). Let $\phi: G \rightarrow H$ be a morphism of groups, then $\operatorname{im} \phi \subseteq H$ is a subgroup of $H$.
Proof. Let $h \in \operatorname{im} \phi$ be any element and consider $g \in \phi^{-1}(h)$, from Proposition 7.3.6 we see that $\phi\left(g^{-1}\right)=\phi(g)^{-1}=h^{-1} \in \operatorname{im} \phi$, thus $\operatorname{im} \phi$ is closed under inverses. Moreover, given another $h^{\prime} \in \operatorname{im} \phi$, there exists $g^{\prime} \in \phi^{-1}\left(h^{\prime}\right)$ and $\phi\left(g g^{\prime}\right)=\phi(g) \phi\left(g^{\prime}\right)=h h^{\prime} \in \operatorname{im} \phi$ - hence $\phi$ is closed under products.

Definition 7.3.9 (Kernel). We define the kernel of a morphism of groups $\phi \in \operatorname{Mor}_{G r p}(G, H)$ as the collection $\operatorname{ker} \phi=\left\{g \in G: \phi(g)=e_{H}\right\}$.

Lemma 7.3.10 (Kernel subgroup). Let $\phi: G \rightarrow H$ be a group morphism. The kernel $\operatorname{ker} \phi \subseteq G$ is a subgroup of $G$.
Proof. Let $g \in \operatorname{ker} \phi$ be any element, then since $\phi\left(g^{-1}\right)=\phi(g)^{-1}=e_{H}^{-1}=e_{H}$, then $g^{-1} \in \operatorname{ker} \phi$. Also, if $u \in \operatorname{ker} \phi$ is another element, then $\phi(g u)=\phi(g) \phi(u)=e_{H} e_{H}=e_{H}$ and hence $g u \in \operatorname{ker} \phi$ - thus $\operatorname{ker} \phi$ is a subgroup of $G$.

Proposition 7.3.11 (Monomorphisms, kernels and injectivity). Let $\phi: G \rightarrow H$ be a morphism of groups. Then the following properties are equivalent:
(a) $\phi$ is a monomorphism in Grp.
(b) $\operatorname{ker} \phi=\left\{e_{G}\right\}$.
(c) $\phi$ is injective in Set.

Proof. - (a) $\Rightarrow$ (b). Suppose $\phi$ is a monormophism and consider the following commutative diagram

$$
\operatorname{ker} \phi \xrightarrow[e]{\stackrel{\iota}{\longrightarrow}} G \stackrel{\phi}{\longrightarrow} H
$$

From the monomorphism definition we find $\iota=e$ and therefore $\operatorname{ker} \phi=\left\{e_{G}\right\}$.

- (b) $\Rightarrow$ (c). If $\operatorname{ker} \phi=\left\{e_{G}\right\}$, then, given elements $g, g^{\prime} \in G$ such that $\phi(g)=\phi\left(g^{\prime}\right)$, then, multiplying both sides by $\phi\left(g^{\prime}\right)^{-1}$ we get

$$
\phi(g) \phi\left(g^{\prime}\right)^{-1}=\phi(g) \phi\left(g^{\prime-1}\right)=\phi\left(g g^{\prime-1}\right)=e_{G}
$$

That is, $g g^{\prime-1} \in \operatorname{ker} \phi$, but since ker $\phi$ is trivial, then $g g^{\prime-1}=e_{G}$ and hence $g=g^{\prime}$ - in other words, $\phi$ is injective.

- $(c) \Rightarrow(a)$. Let $\phi$ be injective - that is, for any pair of set-functions $f, g: A \rightrightarrows G$ from a set $A$, then $\phi f=\phi g$ if and only if $f=g$. In particular, if we impose a group structure in the set $A$ and that both $f$ and $g$ preserve the group structure of $A$, we find that the injectivity in Set implies in the monomorphism $\phi$ as a morphism in Grp.

Proposition 7.3.12. Let $G$ be a group and $H$ be an abelian group. If there exists an injective group morphism $t: G \mapsto H$, then $G$ is abelian.

Proof. Let $x, y \in G$ be any pair of elements, then $\phi(x y)=\phi(x) \phi(y)=\phi(y) \phi(x)=$ $\phi(y x)$, thus $x y=y x$ - since $\phi$ is injective.

The following proposition is a trivial one, but it is a really useful tool to prove the non-existence of non-trivial morphisms between certain groups - it relies on arguments based on the order of elements of each group.

Proposition 7.3.13 (Morphisms and orders). Let $G \in G r p$ be a group admitting an element $g \in G$ of finite order $|g| \in \mathbf{N}$. Let $H$ be any group and consider the morphism $\phi: G \rightarrow H$. We have that $|\phi(g)|$ divides the order of $|g|$.

Proof. Notice that $\phi(g)^{|g|}=\phi\left(g^{|g|}\right)=\phi\left(e_{G}\right)=e_{H}$, hence $|g|$ is a multiple of the order of $\phi(g) \in H$.

Example 7.3.14. Consider for example the collection of morphisms $\operatorname{Mor}_{G r p}\left(C_{7}, C_{15}\right)$. Let $\phi$ be any such morphism. Consider any element $g \in C_{7}$ and recall that $|g|$ must be a divisor of 7 - see Proposition 7.2.20. On the other hand, if $h \in C_{15}$ then $|h|$ divides 15. From Proposition 7.3 .13 we see that if $\phi(g)=h$, then $|h|$ must divide $|g|$, but $\operatorname{gcd}(7,15)=1$, hence $|h|=1$ - that is $h=e_{H}$ and $\phi$ is the trivial morphism $\phi(g)=e_{H}$ for all $g \in C_{7}$. This shows that there is no non-trivial morphism between $C_{7}$ and $C_{15}$.

## Isomorphism of Groups

Proposition 7.3.15 (Isomorphisms are bijections). Let $\phi \in \operatorname{Mor}(\operatorname{Grp})$. Then $\phi$ is an isomorphism if and only if it is a bijection.

Proof. Consider an isomorphism $\phi: G \rightarrow H$. Using the forgetful functor $F: G r p \rightarrow$ Set we see that $F \phi$ is a bijection of sets in Set - recall Lemma 1.4.11 - thus $\phi$ defines a bijection between the elements of $G$ and $H$.

On the other hand, if $\phi$ is a bijection, consider its set-function inverse $(F \phi)^{-1}: H \rightarrow$ $G$. We now show that $(F \phi)^{-1}$ preserves the structures of groups. Since $\phi\left(e_{G}\right)=e_{H}$, then $(F \phi)^{-1}\left(e_{H}\right)=e_{G}$. Moreover, for any $h, h^{\prime} \in H$ - since $\phi$ is surjective - take elements $g, g^{\prime} \in G$ such that $\phi(g)=h$ and $\phi\left(g^{\prime}\right)=h^{\prime}$, then we find that $(F \phi)^{-1}\left(h h^{\prime}\right)=g g^{\prime}=$ $(F \phi)^{-1}(h) \cdot(F \phi)^{-1}\left(h^{\prime}\right)$. This implies in the existence of a naturally induced morphism of groups $\phi^{-1}: H \rightarrow G$ defined by $\phi^{-1}(h)=(F \phi)^{-1}(h)$. It is clear that $\phi^{-1}$ is the right and left inverse of $\phi$, thus $\phi^{-1}$ is the inverse of $\phi$ in Grp and $\phi$ is an isomorphism.

Definition 7.3.16 (Embedding). Let $\phi: G \xrightarrow{\simeq} H$ be an isomorphism of groups. We define the group $\operatorname{im} \phi \subseteq H$ as an embedding of $G$ on $H$.

Proposition 7.3.17. Let $\phi: G \xrightarrow{\simeq} H$ be an isomorphism of groups. Then:

- For all $g \in G$, we have $|g|=|\phi(g)|$.
- $G$ is commutative if and only if $H$ is commutative.

Proof. Let $g \in G$ be any element, then from Proposition 7.3.13 we have that $|\phi(g)|$ divides $|g|$. Since $\phi^{-1}$ exists and is a morphism of groups, it also follows that $|g|$ divides $|\phi(g)|$ — hence $|g|=|\phi(g)|$.

Let $G$ be a commutative group and $G \simeq H$. Let $h, h^{\prime} \in H$ be any elements and consider $\phi^{-1}(h)=g$ and $\phi^{-1}\left(h^{\prime}\right)=g^{\prime}$. From the structure preserving property of $\phi$ we have

$$
h h^{\prime}=\phi(g) \phi\left(g^{\prime}\right)=\phi\left(g g^{\prime}\right)=\phi\left(g^{\prime} g\right)=\phi\left(g^{\prime}\right) \phi(g)=h^{\prime} h
$$

That is, $H$ is commutative. The counter-implication is equivalent and will be omitted.

Lemma 7.3.18 (Inner automorphism). Let $G \in G r p$. For each $g \in G$, the map $\gamma_{g}: G \rightarrow G$ given by $\gamma_{g}(a)=g a g^{-1}$ is an automorphism - called inner automorphism of $G$.

Proof. Let $g \in G$ be any element. Suppose $a \in \operatorname{ker} \gamma_{g}$, then $\gamma_{g}(a)=g a g^{-1}=e_{G}$, hence, $a=g^{-1} e_{G} g=g^{-1} g=e_{G}$, that is $\operatorname{ker} \gamma_{g}=e_{G}-\gamma_{g}$ is injective. Let $g^{\prime} \in G$ be any element, then, $\gamma_{g}\left(g^{-1} g^{\prime} g\right)=g\left(g^{-1} g^{\prime} g\right) g^{-1}=g^{\prime}$, that is, $\gamma_{g}$ is surjective. We conclude that $\gamma_{g}$ is a bijection - hence an isomorphism, so $\gamma_{g} \in \operatorname{Aut}_{G r p}(G)$.

Lemma 7.3.19 (Inner automorphism correspondence). Let $G \in G r p$. The map $\phi: G \rightarrow$ $\operatorname{Aut}_{\mathrm{Grp}}(G)$ defined by the mapping $\phi(g)=\gamma_{g}$ (where $\gamma_{g}$ is defined in Lemma 7.3.18) is a morphism of groups.

Proof. Let $g, g^{\prime} \in G$ be any elements, then

$$
\gamma_{g g^{\prime}}(a)=\left(g g^{\prime}\right) a\left(g g^{\prime}\right)^{-1}=\left(g g^{\prime}\right) a\left(g^{\prime-1} g^{-1}\right)=g\left(g^{\prime} a g^{\prime-1}\right) g^{-1}=\gamma_{g} \gamma_{g^{\prime}}(a)
$$

This being said, its easy to see that $\phi$ preserves the group structure: $\phi\left(g g^{\prime}\right)=\gamma_{g g^{\prime}}=$ $\gamma_{g} \gamma_{g^{\prime}}=\phi(g) \phi\left(g^{\prime}\right)$. Thus $\phi$ is a morphism of groups.

## More Thoughts On Cyclic Groups

We can now state the definition of a cyclic group in a formal manner, it goes as follows:
Definition 7.3.20 (Cyclic group). A group $G$ is said to be cyclic if $G \simeq \mathbf{Z}$ or $G \simeq \mathbf{Z} / n \mathbf{Z}$ for some $n \in \mathbf{N}$.

Proposition 7.3.21. A finite group of order $n \in \mathbf{N}$ is cyclic if and only if it contains an element of order $n$.

Proof. Let $G$ be a cyclic group of order $n$. Since $G$ is finite, then there exists an isomorphism $\phi: G \xrightarrow{\simeq} \mathbf{Z} / n \mathbf{Z}$. Consider the element $g=\phi^{-1}\left([1]_{n}\right) \in G$, from Proposition 7.3.17 we see that $|g|=n$.

Let $G$ be a finite group of order $n$ and $x \in G$ be such that $|x|=n$. Let $\phi: G \rightarrow$ $\mathbf{Z} / n \mathbf{Z}$ be any morphism of groups sending $x \mapsto[1]_{n}$. Consider the collection $G^{\prime}=$ $\left\{e_{G}, x, x^{2}, \ldots, x^{n-1}\right\} \subseteq G$, and notice that, together with the binary operation of $G, G^{\prime}$ becomes a group of $n$ elements - that is, $G^{\prime}=G$ and every element $g \in G$ can be written as $g=x^{k}$ for some $1 \leqslant k \leqslant n$. This implies in $\phi$ injective - thus a bijection. From Proposition 7.3.15 we see that $\phi$ is an isomorphism $G \simeq \mathbf{Z} / n \mathbf{Z}-G$ is a cyclic group, which finishes our proof.

Proposition 7.3.22. The order of the cyclic group $C_{n}$ is equal to $\phi(n)$, where $\phi$ is the Euler totient function - that is, the number of positive integers less than $n$ that are relatively prime to $n$.

Proof. Let $x \in C_{n}$ be a generator of the group - that is, $x^{n}=e$ - then for all $d<n$ such that $\operatorname{gcd}(d, n)=1$ we have $x^{d} \neq x$ and $\left(x^{d}\right)^{n}=e$ thus $x^{d}$ is a generator of $C_{n}$. Therefore the number of distinct elements of $C_{n}$ is the same as the number of positive integers coprime of $n$.

Proposition 7.3.23 (Subgroup). Any subgroup of a cyclic group is cyclic.
Proof. Let $G=\langle g\rangle$ be a cyclic group and $H \subseteq G$ be any subgroup of $G$. If $h \in H$ then there exists $n \in \mathbf{Z}$ such that $h=g^{n}$, thus $g^{n} \in H$ - but then, given any other $h^{\prime} \in H$, there must exist $m \in \mathbf{Z}$ such that $h^{\prime}=g^{n+m}=h^{m}$, thus $H=\langle h\rangle$ is cyclic.

Lemma 7.3.24. Given a group $G$, the collection of inner automorphisms, which we'll denote by $\operatorname{Inn}(G)$, is a subgroup of $\operatorname{Aut}(G)$. Moreover, the following statements are equivalent:
(a) $\operatorname{Inn}(G)$ is cyclic.
(b) $\operatorname{Inn}(G)$ is trivial.
(c) $G$ is abelian.

Therefore, if $\operatorname{Aut}(G)$ is cyclic, the group $G$ is abelian.
Proof. From Lemma 7.3.18 and Lemma 7.3.19 we find that $\operatorname{Inn}(G) \subseteq \operatorname{Aut}(G)$ is indeed a subgroup. Now we prove the equivalences:

- (a) $\Leftrightarrow(b)$. If $\operatorname{Inn}(G)$ is cyclic, then there must exist $g \in G$ for which $\operatorname{Inn}(G)=\left\langle\gamma_{g}\right\rangle$. Therefore, given any $h \in G$, there exists $n \in \mathbf{Z}$ for which $\gamma_{h}=\gamma_{g}^{n}$ - therefore $h g h^{-1}=g^{n} g g^{-n}=g$, thus $h g=g h$. We can thus conclude that $\gamma_{g}$ is the identity morphism $h \mapsto h$, which implies in $\operatorname{Inn}(G)$ being trivial. Now, if we assume $\operatorname{Inn}(G)$ to be trivial by hypothesis, then clearly $\operatorname{Inn}(G)$ is cyclic.
- (b) $\Leftrightarrow$ (c). If $\operatorname{Inn}(G)$ is trivial, then given any pair of elements $g, h \in G$ we have $\gamma_{g}=$ id and thus $g h g^{-1}=h$, which implies in $g h=h g$, thus $G$ is commutative. Now, if $G$ is commutative by hypothesis, we find that $g h g^{-1}=\left(g g^{-1}\right) h=h$ thus $\operatorname{Inn}(G)$ is trivial.

Thus, if $\operatorname{Aut}(G)$ is a cyclic group, we find that the subgroup $\operatorname{Inn}(G) \subseteq \operatorname{Aut}(G)$ is also cyclic (from Proposition 7.3.23), hence $G$ is abelian.

## Some Matrix Groups

Example 7.3.25 (Important matrix groups). Let $k$ be a field and $\operatorname{Mat}_{n \times n}(k)=$ End $_{\text {Vect }_{k}}\left(k^{n}\right)$ be the collection of all $n \times n$ matrices over $k$, for any $n \in \mathbf{Z}_{>0}$. We define the following important groups of square matrices:

1. General linear group: $\mathrm{GL}_{n}(k):=\left\{T \in \operatorname{Mat}_{n \times n}(k): \operatorname{det} T \neq 0\right\}$, the group of invertible matrices.
2. Special linear group: $\mathrm{SL}_{n}(k):=\left\{T \in \mathrm{GL}_{n}(k): \operatorname{det} T=1\right\}$.
3. Orthogonal group: $\mathrm{O}_{n}(k):=\left\{T \in \mathrm{GL}_{n}(k): T T^{*}=T^{*} T=\mathrm{id}_{n}\right\}^{1}$.
4. Special orthogonal group: $\mathrm{SO}_{n}(k):=\left\{T \in \mathrm{O}_{n}(k): \operatorname{det} T=1\right\}$.
5. Unitary group: $\mathrm{U}_{n}(\mathbf{C}):=\left\{T \in \mathrm{GL}_{n}(\mathbf{C}): T T^{\dagger}=T^{\dagger} T=\mathrm{id}_{n}\right\}^{2}$.
6. Special unitary group: $\mathrm{SU}_{n}(\mathbf{C}):=\left\{T \in \mathrm{U}_{n}(\mathbf{C}): \operatorname{det} T=1\right\}$.
7. Lie algebra of the general linear group: $\mathfrak{g l}_{n}(k) \subseteq \mathrm{GL}_{n}(k)$ such that

$$
[T, M]:=T M-M T \in \mathfrak{g l}_{n}(k)
$$

for all $T, M \in \operatorname{gl}_{n}(k)$.
8. Lie algebra of the special linear group: $\mathfrak{s l}_{n}(k):=\left\{T \in \mathfrak{g l}_{n}(k): \operatorname{tr} T=0\right\}$.
9. Lie algebra of the orthogonal group: $\mathfrak{o}_{n}(k):=\left\{T \in \mathfrak{g l}_{n}(k): T+T^{*}=0\right\}$
10. Lie algebra of the unitary group: $\mathfrak{u}_{n}(\mathbf{C}):=\left\{T \in \mathfrak{g l}_{n}(\mathbf{C}): T+T^{+}=0\right\}$.
11. Lie algebra of the special unitary group: $\mathfrak{s u}_{n}(\mathbf{C}):=\left\{T \in \mathfrak{g l}_{n}(\mathbf{C}): \operatorname{tr} T=0\right\}$.

The proof that such examples are indeed groups come immediately from Proposition 6.6.4 and Definition 5.7.6. Notice that all of the above groups are subgroups of the general linear group $\mathrm{GL}_{n}(k)$. It should be noted, however, that the Lie groups presented do not form a group under multiplication.

[^8]Example 7.3.26 (Upper triangular matrices). The collection of upper triangular matrices over a field $k$ is a subgroup of $\mathrm{GL}_{n}(k)$. Let $A$ be any upper triangular matrix, then from definition $\left[a_{i j}\right]_{1 \leqslant i, j \leqslant n}$ is such that $a_{i j}=0$ for all $i>j$ and $\prod_{j=1}^{n} a_{j j} \neq 0$ — therefore, any permutation $\sigma \in S_{n}$ other than the identity will have at least one element $i_{0}>j$. Hence

$$
\operatorname{det} A=\sum_{\sigma \in S_{n}} \operatorname{sign}(\sigma) \prod_{j=1}^{n} a_{\sigma(j) j}=\prod_{j=1}^{n} a_{j j} \neq 0,
$$

which implies in $A \in \mathrm{GL}_{n}(k)$. Moreover, since the sum of two upper triangular matrices is upper triangular, clearly the sum is in $\mathrm{GL}_{n}(k)$ - which proves that the collection of upper triangular matrices is a subgroup, since the existence of inverses is immediate.

Example 7.3.27. Notice that any matrix of the form

$$
A:=\left[\begin{array}{cc}
a+b i & c+d i  \tag{7.4}\\
-c+d i & a-b i
\end{array}\right]
$$

Is such that

$$
\begin{aligned}
A A^{+} & =\left[\begin{array}{cc}
a+b i & c+d i \\
-c+d i & a-b i
\end{array}\right]\left[\begin{array}{cc}
a-b i & -c-d i \\
c-d i & a+b i
\end{array}\right]=\left[\begin{array}{cc}
a^{2}+b^{2}+c^{2}+d^{2} & 0 \\
0 & a^{2}+b^{2}+c^{2}+d^{2}
\end{array}\right], \\
A^{+} A & =\left[\begin{array}{cc}
a-b i & -c-d i \\
c-d i & a+b i
\end{array}\right]\left[\begin{array}{cc}
a+b i & c+d i \\
-c+d i & a-b i
\end{array}\right]=\left[\begin{array}{cc}
a^{2}+b^{2}+c^{2}+d^{2} & 0 \\
0 & a^{2}+b^{2}+c^{2}+d^{2}
\end{array}\right] .
\end{aligned}
$$

Therefore, imposing $a^{2}+b^{2}+c^{2}+d^{2}=1$ in Eq. (7.4) makes $A$ into a matrix of the group $\mathrm{SL}_{2}(\mathbf{C})$. Furthermore, if $T \in \mathrm{GL}_{n}(\mathbf{C})$ is a matrix with a form other than that of Eq. (7.4), then it cannot be the case that $\operatorname{det} T$ equal 1 - thus, every $\mathrm{SL}_{2}(\mathrm{C})$ matrix has the form of $A$.

Proposition 7.3.28. The group $\mathrm{SL}_{2}(\mathbf{Z})$ is generated by the pair matrices

$$
A:=\left[\begin{array}{cc}
0 & -1 \\
1 & 0
\end{array}\right] \quad \text { and } \quad B:=\left[\begin{array}{ll}
1 & 1 \\
0 & 1
\end{array}\right]
$$

Proof. It should be noted that $B^{n}=\left[\begin{array}{ll}1 & n \\ 0 & 1\end{array}\right]$ and $A^{2}=\left[\begin{array}{cc}-1 & 0 \\ 0 & -1\end{array}\right]$ — thus, given any matrix $X=\left[\begin{array}{ll}a & b \\ c & d\end{array}\right] \in \mathrm{SL}_{2}(\mathbf{Z})$ we have

$$
A X=\left[\begin{array}{cc}
-c & -d \\
a & b
\end{array}\right] \quad \text { and } \quad B^{n} X=\left[\begin{array}{cc}
a+n c & b+n d \\
c & d
\end{array}\right]
$$

If we assume that $c \neq 0$, and $|a| \geqslant|c|$ (otherwise, simply multiply from the left by $A$ ), we can apply the division algorithm to find $q_{0}, r_{0} \in \mathbf{Z}$ for which $a=q_{0} c+r_{0}$, satisfying $0 \leqslant r_{0} \leqslant|c|$. Then we can apply

$$
A B^{-q_{0}} X=A\left[\begin{array}{cc}
r_{0} & b-q_{0} d \\
c & d
\end{array}\right]=\left[\begin{array}{cc}
-c & -d \\
r_{0} & b-q_{0} d
\end{array}\right]:=X_{0}
$$

Now, if $r_{0} \neq 0$, we can recursively apply the division algorithm (which I'll carry again just for the sake of clarity): we find $q_{1}, r_{1} \in \mathbf{Z}$ such that $-c=q_{1} r_{0}+r_{1}$, where $0 \leqslant r_{1} \leqslant\left|r_{0}\right|$ - we thus find a new power $q_{1}$ for which

$$
A B^{-q_{1}} X_{0}=\left[\begin{array}{cc}
-r_{0} & q_{0} d-b \\
r_{1} & \left(q_{0} q_{1}-1\right) d-q_{1} b
\end{array}\right]=X_{1} .
$$

This process is ensured to terminate at some point with a zero lower left entry. Since the group is acting on the left of $\mathrm{SL}_{2}(\mathbf{Z})$ and the determinant is always zero, the resulting matrix will be of the form $\left[\begin{array}{ccc} \pm 1 & m \\ 0 & \pm 1\end{array}\right] \in \mathrm{SL}_{2}(\mathbf{Z})$ - which equals either $B^{m}$ or $-B^{-m}$. We conclude that there must exist some $g \in\langle A, B\rangle$ such that $g X= \pm B^{t}$ for some $t \in \mathbf{Z}$ and since $A^{2}=-I_{2}$, we obtain $X= \pm g^{-1} B^{t} \in \mathrm{SL}_{2}(\mathbf{Z})$ and $\pm g^{-1} B^{t} \in\langle A, B\rangle$.

Proposition 7.3.29 (Union of subgroups). Let $G$ be any group, then:
(a) Given subgroups $H, Q \subseteq G$, the union $H \cup Q$ is a subgroup of $G$ if and only if either $H \subseteq Q$ or $Q \subseteq H$.
(b) If $H_{0} \subseteq H_{1} \subseteq \ldots$ is a collection of subgroups of $G$, then the union $\bigcup_{j \geqslant 0} H_{j} \subseteq G$ is a subgroup of $G$.

Proof. (a) If $H \subseteq Q$ or $Q \subseteq H$, then $H \cup Q$ is clearly a subgroup. On the other hand, if $H \cup Q$ is a subgroup, then for every $h \in H$ and $q \in Q$ we must have $h q \in H \cup Q$ which implies that $h q \in H$ or $h q \in Q$, for the first case, we have $h^{-1}(h q)=q \in H$ thus $Q \subseteq H$, for the second case, $(h q) q^{-1}=h \in Q$ implying in $H \subseteq Q$.
(b) Since the composition of elements is only defined for finitely many elements, the proposition follows immediately.

## Group Products

Let $\left(G, \cdot{ }_{G}\right),(H, \cdot H) \in \operatorname{Grp}$ be any objects. We define a binary operation $\cdot:(G \times H)^{2} \rightarrow$ $G \times H$ as the mapping

$$
\begin{equation*}
(g, h) \cdot\left(g^{\prime}, h^{\prime}\right)=\left(g \cdot G g^{\prime}, h \cdot H h^{\prime}\right) \tag{7.5}
\end{equation*}
$$

Such binary operation defines a group structure on $G \times H$. Notice that, given an element $(g, h) \in G \times H$, there exists an element $\left(g^{-1}, h^{-1}\right) \in G \times H$ such that $(g, h) \cdot\left(g^{-1}, h^{-1}\right)=$ $\left(e_{G}, e_{H}\right)$. Moreover, clearly $\left(e_{G}, e_{H}\right) \in G \times H$ is the identity element of the structure. Hence $(G \times H, \cdot) \in G r p$.

Also, the natural projections $\pi_{G}: G \times H \rightarrow G$ and $\pi_{H}: G \times H \rightarrow H$ define morphisms of groups.

Definition 7.3.30 (Direct product). Let $\left\{G_{j}\right\}_{j \in J}$ be a collection of groups. We define the direct product of this family as the group $\prod_{j \in J} G_{j}$ given by elements $\left(x_{j}\right)_{j \in J}$ such that $x_{j} \in G_{j}$. The composition of elements of the direct product is defined component-wise, that is, if $\left(x_{j}\right)_{j \in J},\left(y_{j}\right)_{j \in J} \in \prod_{j \in J} G_{j}$, then $\left(x_{j}\right)_{j \in J}\left(y_{j}\right)_{j \in J}:=\left(x_{j} y_{j}\right)_{j \in J}$. Moreover, inverses are also defined component-wise, $\left(x_{j}\right)_{j \in J}^{-1}:=\left(x_{j}^{-1}\right)_{j \in J}$.

Proposition 7.3.31. The direct products of groups are products on the category of groups, Grp. That is, for all group $W$ and group morphisms $f \in \operatorname{Mor}_{G r p}(W, G)$ and $g \in \operatorname{Mor}_{\mathrm{Grp}}(W, H)$, there exists a unique morphism $\varphi \in \operatorname{Mor}_{\mathrm{Grp}}(W, G \times H)$ such that the following diagram commutes


Proof. We just take $\varphi: W \rightarrow G \times H$ as the mapping $w \stackrel{\varphi}{\mapsto}(f(w), g(w))$. We show that $\varphi$ exists in Grp: let $x, y \in W$ be any elements, then, since $f$ and $g$ are group morphisms, we find that

$$
\varphi(x y)=(f(x y), g(x y))=(f(x) f(y), g(x) g(y))=(f(x), g(x))(f(y), g(y))=\varphi(x) \varphi(y)
$$

That is, $\varphi$ is a group morphism. The uniqueness comes from the covariant functor $F: \operatorname{Grp} \rightarrow$ Set, since Set allows for products and hence the set-function $F \varphi$ is unique.

Remark 7.3.32. We now show that if $G, H \in G r p$ are such that $G \simeq H \times G$, it does not follow that $H$ is trivial. Let $G:=\bigoplus_{j=0}^{\infty} \mathbf{Z}$ and $H:=\mathbf{Z}$ be abelian groups under the natural structure of addition. Notice that $G \simeq H \times G$ by the natural assignment of each element of $G$ to itself in $H \times G$. Since $H$ is non-trivial, we found a counterexample.

Proposition 7.3.33. Let $G$ be a group and $H, Q \subseteq G$ be subgroups for which $H \cap Q=e$ and $H Q=G$ - that is, for every $g \in G$, there exists $h \in H$ and $q \in Q$ such that $g=h q$ - we also impose that $h q=q h$ for every $h \in H$ and $q \in Q$. Then, the morphism of groups $H \times Q \stackrel{\simeq}{\leftrightharpoons} G$ defined by the mapping $(h, q) \mapsto h q$ is an isomorphism.

Proof. Notice that $(h, q)\left(h^{\prime}, q^{\prime}\right)=\left(h h^{\prime}, q q^{\prime}\right) \mapsto\left(h h^{\prime}\right)\left(q q^{\prime}\right)=(h q)\left(h^{\prime} q^{\prime}\right)$ thus the map is indeed a morphism of groups. Moreover, since every element of $G$ can be written as a product $H Q$, it follows that the map is surjective. Now, let $(h, q)$ be in the kernel of the morphism, then $h q=e$ which implies in $h=q^{-1}$ but then $h \in H \cap Q$ and by hypothesis $h=e$ - thus the morphism is injective.

### 7.4 Quotient Groups - The Birth of Normal Subgroups

## Cosets

Definition 7.4.1 (Coset). Let $G$ be a group and $H$ be a subgroup of $G$. Given any $g \in G$, a left coset of $H$ in $G$ induced by $g$ and denoted by $g H$ is a set whose elements have the form $g h$ for each $h \in H$. A right coset of $H$ in $G$ induced by $g$ is denoted $H g$ and is a set consisting of elements of the form $h g$ for each $h \in H$. An element of a coset is commonly called coset representative.

Corollary 7.4.2. Let $G$ be a group and $H$ be subgroup of $G$, then, for every $h \in H$, we have

$$
h H=H h=H .
$$

Proof. Notice that, given $x \in H$, the element $h\left(h^{-1} x\right)=x \in h H$, thus $H \subseteq h H$, on the other hand, it's clear that $h H \subseteq H$, since $h H$ is composed of product of elements of $H$, which itself is closed under products. The same analogous proof goes for Hh so I won't bother to write it down.

Corollary 7.4.3 (Equal cosets). Let $G$ be a group and $H \subseteq G$ be a subgroup. Given $x, y \in G$, if the cosets $x H$ and $y H$ share any common element, then $x H=y H$.

Proof. Let $g \in x H \cap y H k$, then there exists $x h \in x H$ and $y h^{\prime} \in y H$ such that $x h=g=$ $y h^{\prime}$, then, in particular, $x=y h^{\prime} h^{-1}$ moreover, since $H$ is a subgroup, it is clear that $h^{\prime} h^{-1} \in H$ then $x H=\left(y h^{\prime} h^{-1}\right) H=y\left(h^{\prime} h^{-1}\right) H=y H$.

Definition 7.4.4 (Index). Let $G$ be a group and $H \subseteq G$ be a subgroup. The number of left cosets of $H$ in $G$ is denoted by [ $G: H$ ], which will be commonly referred to as the index of $H$ in $G$.

Corollary 7.4.5. If we denote by * the trivial group, the order of a group $G$ is the same as $[G: *]$ - that is, $|G|=[G: *]$.

Proof. One can view the trivial group * as a subgroup of $G$ containing only the identity. Notice that the number of left cosets of $*$ will be exactly the number of elements of $G$, that is $[G: *]=|G|$.

Proposition 7.4.6. Let $G$ be a group, and $H \subseteq G$ be a subgroup, and $Q \subseteq H$ be a subgroup. Then, if any two of the quantities $\{[G: H],[H: Q],[G: Q]\}$ is finite, the third is also finite and the following equality holds

$$
[G: H][H: Q]=[G: Q] .
$$

Proof. Let $\left\{x_{i}\right\}_{i \in I} \subseteq H$ be coset representatives of $Q$, and $\left\{y_{j}\right\}_{j \in J} \subseteq G$ be coset representatives of $H$ - that is, each one of the collections $\left\{x_{i} Q\right\}_{i \in I}$ and $\left\{y_{j} H\right\}_{j \in J}$ have pairwise disjoint elements, and $H=\bigcup_{i \in I} x_{i} Q$, and $G=\bigcup_{j \in J} y_{j} H$. Then we have that $G=\bigcup_{(i, j) \in I \times J} y_{j} x_{i} Q$, and our goal will be to prove that $y_{j} x_{i} Q \cap y_{i^{\prime}} x_{j^{\prime}} Q=\varnothing$. Suppose on the contrary that their intersection is non-empty, which by Corollary 7.4.3 implies $y_{j} x_{i} Q=y_{j^{\prime}} x_{i^{\prime}} Q$. Since $x_{j}, x_{j^{\prime}} \in H$, we have $y_{j} x_{i} Q H=y_{j} x_{i} H=y_{j} H$ and analogously $y_{j^{\prime}} x_{i^{\prime}} Q H=y_{j^{\prime}} H$ - thus $y_{j} H=y_{j^{\prime}} H$, which implies in $y_{j}=y_{j^{\prime}}$. This shows that the collection $\left\{y_{j} x_{i}\right\}_{(i, j) \in I \times J} \subseteq G$ are coset representatives for $Q$ and therefore $[G: H][H: Q]=[G: Q]$.

Corollary 7.4.7 (Lagrange's theorem). Let $G$ be a finite group, then the order of any subgroup $H$ of $G$ divides the order $|G|$.

Proof. Since $|G|=[G: *]$ is finite, any subgroup $H$ of $G$ is also finite and therefore $[G: H][H: *]=[G: *]$, which is exactly the same as $[G: H]|H|=|G|$.

Example 7.4.8 (Prime order). Let $G$ be a group with order $|G|:=p$ prime. Choose any $g \in G$ with $g \neq e$, and consider the subgroup $H:=\langle g\rangle$. From Proposition 7.4 .6 we find that $[G: H]|H|=p$, hence $|H|$ divides $p$, but since $|H| \leqslant p$, then $|H|=p$ and therefore $H=G$. This implies that any non-identity element of $G$ generates the whole group, which is the same as to say that $G$ is cyclic.

## Normal Subgroups

Definition 7.4.9 (Normal subgroup). Let $G$ be a group. We define a normal subgroup to be the kernel of some morphism of groups in $\operatorname{Mor}_{G r p}(G,-)^{3}$. In other words, a subgroup $N \subseteq G$ is normal if there exists a morphism of groups $\phi: G \rightarrow H$, for some group $H$, for which $\operatorname{ker} \phi=N$.

Definition 7.4.10 (Quotient group). Let $G$ be a group and $N$ be a normal subgroup of $G$. We denote by $G / N$ the collection of all left cosets of $N$ in $G$, on the other hand, $G \backslash N$ denotes the collection of all right cosets of $N$ in $G$. Moreover, we view $G / N$ (and $G \backslash N)$ as groups where:

- The product of two cosets $x N$ and $y N$ (or, respectively, $N x$ and $N y$ ) is given by $(x N)(y N):=(x y) N$ which is again a left coset in $G / N$ (conversely, $(N x)(N y):=$ $N(x y) \in G \backslash N)$.
- Given any coset $x N$ (respectively, $N x$ ), its inverse is given by $x^{-1} N$ (respectively, $N x^{-1}$ ).
- The identity of the group is $N$.

The group $G / N$ is commonly referred to as the quotient group of $G$ by $H$.
Proposition 7.4.11. Let $G$ be a group. A subgroup $N \subseteq G$ is normal if and only if, for every $g \in G$, we have $g N g^{-1}=N$.

Proof. First, suppose that $N$ is a normal group and $\phi: G \rightarrow H$ is a morphism such that $\operatorname{ker} \phi=N$, then if $g \in G$ is any element, we see that for any $n \in N$ we have $\phi(g n)=\phi(g) \phi(n)=\phi(g)$ and analogously $\phi(n g)=\phi(n) \phi(g)=\phi(g)$, thus in general $g N=N g=\phi^{-1}(\phi(g))$. Note that if we multiply both groups on the right by $g^{-1}$ we get $g N g^{-1}=N$, as wanted.

On the other hand, let $N \subseteq G$ be a subgroup such that $g N g^{-1}=N$ for every $g \in G$, then in particular left cosets are equal to right cosets because, multiplying on the right by $g$ we obtain $g N=N g$. Consider the group of left cosets $G / N$ (since right and left cosets are equivalent in this specific case, we could also have considered $G \backslash N$ ) and define the morphism of groups $\pi: G \rightarrow G / N$ by the mapping $\pi(g)=g N$. Notice that if $g \in \operatorname{ker} \pi$, then $g N=N$, which implies that $g \in N$, moreover, if $n \in N$ is any element, then $\pi(n)=n N=N$ - thus ker $\pi=N$.

[^9]In fact the morphism $\pi: G \rightarrow G / N$, defined above by the map $g \stackrel{\psi}{\mapsto} g N$, is so important we are even going to distinguishably call it the canonical projection map of $G$ onto the factor group $G / N$. It is trivial that such canonical projection $\pi$ is surjective.

Corollary 7.4.12. Let $G$ be a group and $H \subseteq G$ be a subgroup. Then $H$ is normal in $G$ if and only if, for all given $\gamma \in \operatorname{Inn}(G)$, we have $\gamma(H) \subseteq H$. Therefore, the morphism $\operatorname{Inn}(G) \rightarrow \operatorname{Aut}(H)$ mapping $\left.\gamma_{g} \mapsto \gamma_{g}\right|_{H}$ is well defined.

Corollary 7.4.13 (Intersection of normal subgroups is normal). Let $\left\{N_{j}\right\}_{j \in J}$ be any collection of normal subgroups of a given group $G$. Then $N:=\bigcap_{j \in J} N_{j}$ is a normal subgroup of $G$.

Proof. Let $n \in N$ and $g \in G$ be any two elements, then $n \in N_{j}$ for all $j \in J$ and from the normal condition we obtain that $g n g^{-1} \in N_{j}$ for all $j \in J$ as well - which implies in $g n g^{-1} \in N$.

Proposition 7.4.14. If $G$ is a group and $N \subseteq G$ is a subgroup with index $[G: N]=2$, then $N$ is normal in $G$.

Proof. Suppose $N \subseteq G$ is a subgroup with index 2 - that is, $G / N$ consists only of two cosets. Thus there must exist $g \in G$ such that $g N \neq N$ and consequently $N g \neq N$ moreover, since $G=N \cup g N=N \cup N g$, one concludes that $g N=N g$.

Definition 7.4.15 (Normal closure). Given a group $G$ and a subset $S \subseteq G$, the normal closure of $S$ is the subgroup of $G$ generated by all elements of the form $g^{-1} s g$ - for $g \in G$ and $s \in S$.

## Quotient Group Properties

We now study the properties of factorization of maps between groups and sequences of maps from the viewpoint of quotientings. One interesting immediate example is when $H \subseteq G$ is a subgroup of a group $G$, then

$$
H ゝ \xrightarrow{\iota} G \xrightarrow{\pi} G / H
$$

which is a short exact sequence. Moreover, given any groups $G, Q$ and $K$, if we have an exact sequence

$$
* \longrightarrow Q \xrightarrow{f} G \xrightarrow{g} K \longrightarrow *
$$

we can conclude that $f$ is injective, and $g$ is surjective - this comes from the fact that $\operatorname{ker} f=e_{Q}$ and, since the kernel of $K \rightarrow 0$ is the whole group $K$, we necessarily have $g(G)=K$. Moreover, if we define $H:=\operatorname{ker} g$, there exists a natural identification


Proposition 7.4.16 (Universal property of quotient groups). Let $G$ be a group and $N$ be a normal subgroup of $G$. Then, given a group $Q$ together with a morphism of groups $\phi: G \rightarrow Q$ such that $N$ is the kernel of $\phi$, there exists a unique morphism $\phi_{*}: G / N \rightarrow Q$ such that the diagram

is commutative. Moreover, the map $\phi_{*}$ is injective. The morphism $\phi_{*}$ induces an isomorphism of groups $\bar{\phi}: G / N \xrightarrow{\simeq} \operatorname{im} \phi$ and therefore we have the following factorization


Proof. Define $\phi_{*}$ simply as the mapping $g H \rightarrow \phi(g)$, then clearly $\phi_{*} \pi=\phi$. Moreover, given any $x \in G / N$, the fiber $\pi^{-1}(x)$ is non-empty, thus $\phi_{*}$ cannot assume any other values beside those specified by $\phi$, which implies in its uniqueness. Moreover, since $\operatorname{ker} \phi=N$, if $n \in N$ then $\phi_{*} \pi(n)=\phi_{*}(n N)=e_{Q}$ but $n N=N$ thus $N \in \operatorname{ker} \phi_{*}$, moreover, if $x N \in \operatorname{ker} \phi_{*}$ it follows that $\phi(x)=e_{Q}$ then $x \in N$ and therefore $x N=N$ — this implies that $\operatorname{ker} \phi_{*}=N$, which is the identity element of $G / N$, thus $\phi_{*}$ is injective.

Corollary 7.4.17 (First isomorphism). Let $\phi: G \rightarrow H$ be a surjective morphism of groups. There exists a canonical isomorphism of groups

$$
G / \operatorname{ker} \phi \simeq H
$$

Proof. By the universal property, the induced map $\phi_{*}: G / \operatorname{ker} \phi \mapsto H$ is already injective. From hypothesis, $\phi$ being surjective implies that $\phi_{*}$ is surjective since $\phi=\pi \phi_{*}$.

Corollary 7.4.18. Let $G$ be a group and $H$ be a subgroup of $G$. Define $S$ to be the subgroup of $G$ consisting of the intersection of all normal subgroups of $G$ containing $H$. Then $S$ is normal in $G$ and is the smallest normal subgroup of $G$ containing $H$. Let $Q$ be a group and $\phi: G \rightarrow Q$ be a morphism of groups such that $H \subseteq \operatorname{ker} \phi$. Then we have $S \subseteq \operatorname{ker} \phi$, and there exists a unique morphism $\phi_{*}: G / S \mapsto Q$ such that the following diagram commutes

moreover, $\phi_{*}$ is injective.

Proof. As before, we just define $\phi_{*}(x S):=\phi(x)$. The commutativity follow from construction, uniqueness follows from the universal property. Moreover, $S$ is given by the intersection of arbitrarily many normal subgroups of $G$ containing $H$, thus in particular $H \subseteq S$ and, for every $g \in G$, we have $g S g^{-1}=S$ - thus $S$ is indeed normal.

Corollary 7.4.19 (Third isomorphism). Let $G$ be a group and $H \subseteq G$ be a normal subgroup of $G$. Let $Q \subseteq G$ be a subgroup containing $H$. Then, $Q / H$ is normal in $G / H$ if and only if $Q$ is normal in $G$, if that is the case, then we have a canonical isomorphism

$$
\frac{G / H}{Q / H} \simeq G / Q .
$$

Proof. Suppose $Q$ is normal in $G$. Since $H$ is contained in $Q$, and $\operatorname{ker}(G \rightarrow G / Q)=Q$, we can apply the universal property of quotients Corollary 7.4.18 to the subgroup $H$ of the kernel and find a unique induced injective morphism $G / H \mapsto G / Q$ - whose kernel, on the other hand, is $Q / H$, thus $Q / H$ is normal in $G / H$.

For the converse, let $Q / H$ be normal in $G / H$, and consider the morphism given by the composition of the canonical projections

$$
G \longrightarrow G / H \longrightarrow \frac{G / H}{Q / H}
$$

which has a kernel given by $Q$ - thus $Q$ is normal in $G$. With that, using Corollary 7.4.18, the induced map is the wanted canonical isomorphism.

In the context of the last corollary, we can visualize the propositions by the following diagram, which commutes


Corollary 7.4.20 (Second isomorphism). Let $G$ be a group, and let $H$ and $Q$ be subgroups of $G$ such that $H \subseteq N_{G}(Q)$ - that is, $H$ is contained in the normalizer of $Q$. Then we have the following canonical isomorphism

$$
\frac{H}{H \cap Q} \simeq \frac{H Q}{Q}
$$

Proof. Since $H \subseteq N_{G}(Q)$, then $h Q h^{-1}=Q$ for every $h \in H$, so that $Q$ is normal in $H$. Since both $H$ and $Q$ are normal in $H$, their intersection $H \cap Q$ is also normal in $H$. Moreover, $H Q=Q H$ and, given $h q \in H Q$, the element $q^{-1} h^{-1} \in Q H=H Q$ exists, thus $H Q$ is closed under inverses, and clearly closed under products, thus $H Q$ is a subgroup of $G$. Consider now the surjective morphism $H \rightarrow H Q / Q$ given by the mapping $h \mapsto h Q$ - whose kernel is $H \cap Q$, and therefore, by the quotient universal property, there exists a unique injective morphism $H /(H \cap Q) \mapsto H Q / Q$, which is also surjective by the construction of $H \rightarrow H Q / Q$.

Proposition 7.4.21 (Morphism preimage preserve normality). Let $G$ and $H$ be groups and $N \subseteq H$ be a normal subgroup. If $f: G \rightarrow H$ is a morphism, then $f^{-1}(N) \subseteq G$ is a normal subgroup of $G$.
Proof. Let $x \in G$ and $y \in f^{-1}(N)$ be any two element of $G$. We have $f\left(x y x^{-1}\right)=$ $f(x) f(y) f(x)^{-1} \in N$, thus $x y x^{-1} \in f^{-1}(N)$.

Notice that the previous proposition gets us the following commutative diagram


Moreover, since ker $\pi f=f^{-1}(N)$, we can use the universal property to obtain


If we now impose surjectivity to the morphism $f$, we find that $\phi$ must also be surjective, giving rise to an isomorphism $G / f^{-1}(N) \simeq H / N$. The last two diagrams can be encoded in a single commutative diagram:


## Centralizers \& Normalizers

Definition 7.4.22 (Centralizers and normalizers). Let $G$ be a group and $S \subseteq G$ be any set of elements. We define the following groups:
(a) The normalizer of $S$ is defined as the group $N(S):=\left\{g \in G: g S g^{-1}=S\right\}$.
(b) A centralizer of $S$ is defined as the group $Z(S):=\left\{g \in G: g s g^{-1}=s\right.$, for all $\left.s \in S\right\}$. The centralizer $Z(G)$ is commonly called the center of $G$.

The normalizer and centralizer are indeed groups, notice that if $g \in N(S)$, then $g^{-1} \in N(S)$ since $g S g^{-1}=S$ it follows, by multiplying on the left by $g^{-1}$ and on the right by $g$, that $S=g^{-1} S g$. Moreover, given any two elements $g, h \in N(S)$, we have

$$
(g h)^{-1} N(g h)=\left(h^{-1} g^{-1}\right) N(g h)=\left(h^{-1} N h\right)\left(g^{-1} N g\right)=S .
$$

For the case of the centralizer the proof is analogous as the one just made for normalizers, where instead of $S$ we would be considering any element $s \in S$.

In the following lemmas, let $G$ be a group and $H$ be a subgroup of $G$.

Lemma 7.4.23. If $Q \subseteq G$ is any subgroup containing $H$, and $H$ is normal in $Q$, then $Q \subseteq N_{G}(H)$ - that is, the normalizer $N_{G}(H)$ of $H$ is the largest subgroup of $G$ such that $H$ is normal.

Proof. Let $K$ be a group and $\phi: Q \rightarrow K$ be a morphism of groups such that ker $\phi=$ $H$, then given any $q \in Q$, and any $h \in H$, we have $\phi\left(q h q^{-1}\right)=\phi(q) \phi(h) \phi(q)^{-1}=$ $\phi(q) \phi(q)^{-1}=e_{K}$ - thus $q h q^{-1} \in H$ and since $q H q^{-1}=H$ from hypothesis, then $q \in N(H)$. The last statement follows clearly from the proposition.

Lemma 7.4.24. If $Q \subseteq N_{G}(H)$ is a subgroup, then $Q H$ is a group and $H$ is normal in QH.

Proof. First we verify that $Q H$ is indeed a group. Let $q h \in Q H$ be any element. Since $Q \subseteq N_{G}(H)$, then $Q H Q^{-1}=H$ and therefore $Q H=H Q$ - hence there exists $q^{\prime} h^{\prime} \in Q H$ such that $q^{\prime} h^{\prime}=q^{-1} h^{-1}$ so that $Q H$ is closed under inverses. Moreover, given any two elements $q h, q^{\prime} h^{\prime} \in Q H$, we have $(q h)\left(q^{\prime} h^{\prime}\right)=q\left(h q^{\prime}\right) h^{\prime}$, but since there exists $q^{\prime \prime} h^{\prime \prime} \in Q H$ such that $q^{\prime \prime} h^{\prime \prime}=h q^{\prime}$ then $q\left(h q^{\prime}\right) h^{\prime}=q\left(q^{\prime \prime} h^{\prime \prime}\right) h^{\prime}=\left(q q^{\prime \prime}\right)\left(h^{\prime \prime} h^{\prime}\right) \in Q H$ - thus $Q H$ is also closed under products, and therefore $Q H$ is a group.

Let $q h \in Q H$ be any element and consider the group $(q h) H\left(h^{-1} q^{-1}\right)$. Given any element $h^{\prime} \in H$, we have that $\left(q^{-1} h^{-1}\right) h^{\prime}(q h) \in H$, thus $(q h)\left[\left(q^{-1} h^{-1}\right) h^{\prime}(q h)\right]\left(h^{-1} q^{-1}\right)=$ $h^{\prime}$ thus $H \subseteq(q h) H\left(h^{-1} q^{-1}\right)$. Moreover, clearly $(q h) H\left(h^{-1} q^{-1}\right) \subseteq H$ thus the equality holds - which implies that $H$ is normal on the group $Q H$.

## Epimorphisms

We lay out the most difficult part of Proposition 7.4.27 in the following lemma.
Lemma 7.4.25 (Epimorphisms are surjections). In the category of groups, an epimorphism is a surjective set-function.

Proof. Let $\phi: H \rightarrow G$ be an epimorphism in Grp - our goal will be to prove that $\operatorname{im} \phi=G$.

Suppose, for the sake of contradiction, that $[G: \operatorname{im} \phi]=2$ - so that, by Proposition 7.4.14 $\operatorname{im} \phi$ is normal in $G$. Now, if we consider the canonical projection and the trivial morphism $\pi, 0: G \rightrightarrows G / \operatorname{im} \phi$ we see that $\pi \phi=0 \phi$ - however, since $\phi$ is epic, this would imply in $\pi=0$ which cannot possibly be true. Hence [ $G: \operatorname{im} \phi$ ] $\geqslant 3$ necessarily and thus $\operatorname{im} \phi$ is .

Define the set-function $\sigma: G / \operatorname{im} \phi \rightarrow G / \operatorname{im} \phi$ to be a permutation of the cosets of $G / \operatorname{im} \phi$ such that $\sigma(\operatorname{im} \phi):=\operatorname{im} \phi$ is the only fixed point of $\sigma$ - which can occur only because $[G: \operatorname{im} \phi]>2$. We now consider the set-functions given by the canonical projection $\pi: G \rightarrow G / \operatorname{im} \phi$ and a map $\eta: G / \operatorname{im} \phi \rightarrow G$ defined so that $\pi \eta=\mathrm{id}_{G / \mathrm{im} \phi}$ and $\eta(\operatorname{im} \phi):=e_{G}$ - which is possible since $\pi$ is surjective and therefore has a right-inverse.

One can define a set-function $\alpha: G \rightarrow \operatorname{im} \phi$ for which every $g \in G$ can be written uniquely as

$$
\begin{equation*}
g=\alpha(g) \eta \pi(g) \tag{7.6}
\end{equation*}
$$

Since $G$ is a group, the existence of $\alpha$ is ensured - simply define $\alpha(g):=g(\eta \pi(g))^{-1}$. The unicity of $\alpha$ comes from the fact that if $\beta: G \rightarrow \operatorname{im} \phi$ is another map such that $g=\beta(g) \eta \pi(g)$, then by the injectivity of $\eta$ we conclude that $\beta=\alpha$. In the special case of $g \in \operatorname{im} \phi$ then $\eta \pi(g)=\eta(\operatorname{im} \phi)=e_{G}$, and thus $\alpha(g):=g$.

Define $\lambda: G \rightarrow G$ to be the set-function given by $\lambda(g):=\alpha(g) \eta(\sigma \pi(g))$, for all $g \in G$. Notice that $\lambda$ is nothing but a permutation on $G$ since every element of $G$ is uniquely written as in Eq. (7.6) - and $\sigma$ merely permutes the cosets of $G / \operatorname{im} \phi$. Let $P$ be the group of all permutations of $G$ and consider group morphisms $k, \ell: G \rightrightarrows P$ defined by

$$
\begin{align*}
& k(g)(x):=g x, \text { for all } g, x \in G  \tag{7.7}\\
& \ell(g):=\lambda^{-1} k(g) \lambda, \text { for all } g \in G \tag{7.8}
\end{align*}
$$

Indeed, given $g, g^{\prime} \in G$ we have

$$
\begin{gathered}
k\left(g g^{\prime}\right)(x)=\left(g g^{\prime}\right) x=g\left(g^{\prime} x\right)=k(g)\left(k\left(g^{\prime}\right)(x)\right), \\
\ell\left(g g^{\prime}\right)=\lambda^{-1} k\left(g g^{\prime}\right) \lambda=\lambda^{-1} k(g) k\left(g^{\prime}\right) \lambda=\ell(g) \ell\left(g^{\prime}\right)
\end{gathered}
$$

thus both are group morphisms. The condition for $k(g)=\ell(g)$ is the same as $\lambda k(g)=$ $k(g) \lambda$, in turn if $x \in G$ is any element, the requirement that

$$
\begin{equation*}
\lambda(g x)=g \lambda(x) \tag{7.9}
\end{equation*}
$$

is also an equivalent condition.
For the case where $g \in \operatorname{im} \phi$ and $x \in G$ is any element, one has $\pi(g x)=\pi(g) \pi(x)=$ $\pi(x)$ and $\alpha(g x)=g \alpha(x)$. Therefore,

$$
\lambda(g x)=\alpha(g x) \eta(\sigma \pi(g x))=g \alpha(x) \eta(\sigma \pi(x))=g \lambda(x) .
$$

Using Eq. (7.9) one concludes that $k(g)=\ell(g)$ for all $g \in \operatorname{im} \phi$ - which in turn implies in $k \phi=\ell \phi$. Note however that since $\phi$ is an epimorphism of groups by hypothesis, we conclude that $k=\ell$ in general - thus $\lambda(g x)=g \lambda(x)$ is true for all choices of $g, x \in G$. Fix $g_{0} \in G$ and consider the particular case where $x:=e_{G}$, we thus have $\lambda\left(g_{0}\right)=g_{0} \lambda\left(e_{G}\right)=g_{0}$ - therefore, $g_{0}=\lambda\left(g_{0}\right)=\alpha\left(g_{0}\right) \eta\left(\sigma \pi\left(g_{0}\right)\right)$ from the definition of $\lambda$, and $g_{0}=\alpha\left(g_{0}\right) \eta \pi\left(g_{0}\right)$ from Eq. (7.6). Combining both equations for $g_{0}$ we conclude that $\eta\left(\sigma \pi\left(g_{0}\right)\right)=\eta \pi\left(g_{0}\right)$, which in turn implies in $\sigma \pi\left(g_{0}\right)=\pi\left(g_{0}\right)$ since $\eta$ is injective. Furthermore, $\sigma$ was constructed so that im $\phi$ was its only fixed point, that is, $\pi\left(g_{0}\right)=\operatorname{im} \phi$ and therefore $g_{0} \in \operatorname{im} \phi$. This shows that $\operatorname{im} \phi=G$ as wanted.

Remark 7.4.26. The clever reader may be tempted to think that every epimorphism in the category of groups is a split epimorphism - so that the forgetful functor Grp $\rightarrow$ Set still preserves it. However, things are not as bright as one might think.

Consider for instance the natural projective group morphism $\mathbf{Z} / 5 \mathbf{Z} \rightarrow \mathbf{Z} / 2 \mathbf{Z}$ sending $[x]_{5} \mapsto[x]_{2}$. Although an epimorphism, such projection is not a split epimorphism - the lack of elements of order 2 in $\mathbf{Z} / 5 \mathbf{Z}$ does not allow the existence of non-trivial morphisms of groups $\mathbf{Z} / 2 \mathbf{Z} \rightarrow \mathbf{Z} / 5 \mathbf{Z}$.

Proposition 7.4.27 (Epimorphisms in Grp). Let $\phi: G \rightarrow H$ be a group morphism. The following propositions are equivalent:
(a) The morphism $\phi$ is an epimorphism in Grp.
(b) The set function $\phi$ is surjective in Set.

Proof. For $(\mathrm{a}) \Rightarrow(\mathrm{b})$, we have Lemma 7.4.25. On the other hand, the proof that (b) $\Rightarrow$ (a) is straightforward: let $f_{1}, f_{2}: H \rightrightarrows Q$ be any two group morphisms such that $f_{1} \phi=f_{2} \phi$, then given any $h \in H$ there exists $g \in G$ such that $\phi(g)=h$ and hence $f_{1} \phi(g)=f_{1}(h)=f_{2}(h)=f_{2} \phi(g)$ - since $h$ was chosen arbitrarily over $H$, we conclude that $f_{1}=f_{2}$.

## Examples \& Consequences

Proposition 7.4.28 (Modular group $\mathrm{PSL}_{2}(\mathbf{Z})$ ). Let $\sim$ be an equivalence relation on $\mathrm{SL}_{2}(\mathbf{Z})$ for which $A \sim B$ if and only if $A= \pm B$. We define the modular group as the quotient of $\mathrm{SL}_{2}(\mathbf{Z})$ by this equivalence relation, that is

$$
\operatorname{PSL}_{2}(\mathbf{Z}):=\mathrm{SL}_{2}(\mathbf{Z}) / \sim
$$

The modular group so constructed is generated by the cosets of the matrices

$$
\left[\begin{array}{cc}
0 & -1 \\
1 & 0
\end{array}\right] \quad \text { and } \quad\left[\begin{array}{cc}
0 & -1 \\
1 & 1
\end{array}\right] .
$$

Proof. Let's consider the canonical projection morphism $\pi$ : $\mathrm{SL}_{2}(\mathbf{Z}) \rightarrow \operatorname{PSL}_{2}(\mathbf{Z})$ and let $g:=\left[\begin{array}{cc}0 & -1 \\ 1 & 0\end{array}\right]$ and $h:=\left[\begin{array}{ll}1 & 1 \\ 0 & 1\end{array}\right]$, then since $g$ and $g h=\left[\begin{array}{cc}0 & -1 \\ 1 & 1\end{array}\right]$ generate $\operatorname{SL}_{2}(\mathbf{Z})$, the classes $\pi(g)$ and $\pi(g h)$ generate $\mathrm{PSL}_{2}(\mathbf{Z})$. Moreover, since $g^{2}=-I_{2}$, the order of $\pi(g)$ is 2 on the other hand, $(g h)^{3}=-I_{2}$ then $\pi(g h)$ has order 3 .

Proposition 7.4.29. Let $G$ be a finite abelian group. If $p$ is a prime divisor of the order of $G$, then there exists an element of $G$ whose order is $p$.

Proof. Since $p$ divides $|G|$, let $m \in \mathbf{Z}$ such that $|G|=p m$. We proceed via strong induction on $m$.

- The base case $m=1$, one has $|G|=p$. If $g \in G$ is any element, then $\langle g\rangle$ is a subgroup of $G$ and, by Lagrange's theorem, we have that $|\langle g\rangle|$ divides $|G|$ - thus it must be the case that $|\langle g\rangle|=p$. Therefore $g$ is an element of order $p$, since $p$ is prime.
- For the inductive hypothesis, let $m>1$ and assume that that the proposition is true for all $m^{\prime}<m$.
- We now prove the case for $m$, that is, $|G|=p m$. Let $g \in G$ be any element and consider the subgroup $\langle g\rangle$ of $G$.
If $p$ divides the order of $g$, then there exists $a \in \mathbf{Z}$ such that $p a=|g|$, hence the element $a g \in G$ has order $p$ - if this is the case, we are done.

Otherwise, since $G$ is abelian, then $\langle g\rangle$ is a subgroup and we may consider the quotient $G /\langle g\rangle$. By Proposition 7.4.6 we find that $|G /\langle g\rangle|=|G| /|\langle g\rangle|=p m /|g|$. Since $p$ does not divide $|g|$ by hypothesis, then $|g|$ must divide $m$, therefore $m /|g|<m$. By the inductive hypothesis, the proposition is true for positive integers less than $m$, hence there exists $h+\langle g\rangle \in G /\langle g\rangle$ with order $p$. If we consider the natural projection $\pi: G \rightarrow G /\langle g\rangle$, by Proposition 7.3 .13 we have that $|\pi(h)|=|h+\langle g\rangle|=p$ divides the order of $h$. If $|h|=p b$ for some $b \in \mathbf{Z}$, then the element $b h \in G$ has order $p$ and we are done.

Solve Aluffi problems from section 8 (quotient groups)

### 7.5 Group Towers \& Solvability

## Change from "towers" to "series"

Definition 7.5.1 (Tower of subgroups). Let $G$ be a group. A finite sequence of groups $\left(G_{0}, \ldots, G_{n}\right)$ is said to be a tower of subgroups of $G$ if

$$
G=G_{0} \supsetneq G_{1} \supsetneq \cdots \supsetneq G_{n}
$$

The tower may be classified as a normal tower if $G_{j+1}$ is normal in $G_{j}$ for every $0 \leqslant j \leqslant$ $n-1$. The tower is said to be abelian (respectively, cyclic) if it is a normal tower and each quotient $G_{j} / G_{j+1}$ is abelian (respectively, cyclic).

A direct corollary of Proposition 7.4.21 goes as follows.
Corollary 7.5.2. Let $f: G \rightarrow H$ be a group morphism, and $H \supsetneq H_{0} \supsetneq \cdots \supseteq H_{n}$ be a normal tower in $H$. Then, if we define subgroups $G_{j}:=f^{-1}\left(H_{j}\right)$ for every $0 \leqslant j \leqslant n$, the sequence $\left(G_{0}, \ldots, G_{n}\right)$ forms a normal tower in $G$.

Moreover, if $\left(H_{0}, \ldots, H_{n}\right)$ is an abelian tower (respectively, cyclic tower) in $H$, then $\left(G_{0}, \ldots, G_{n}\right)$ is an abelian tower (respectively, cyclic tower) in $G$.

Proof. The first part is clear. For the second, notice that since for all $0 \leqslant j \leqslant n-1$ we have the following commutative diagram

then the injection $G_{j} / G_{j+1} \mapsto H_{j} / H_{j+1}$ allow us to view $G_{j} / G_{j+1}$ as a subgroup of $H_{j} / H_{j+1}$ - thus necessarily abelian (respectively, cyclic).

Definition 7.5.3 (Refinement). Given a tower $G=G_{0} \supsetneq \cdots \supsetneq G_{n}$ of subgroups of $G$, a refinement of such tower is given by the insertion of a finite sequence of subgroups $\left(G_{n+1}, \ldots, G_{m}\right)$ to the end of the tower - that is,

$$
G=G_{0} \supsetneq \cdots \supsetneq G_{n} \supsetneq G_{n+1} \supsetneq \cdots \supsetneq G_{m} .
$$

Definition 7.5.4 (Solvable group). A group $G$ is said to be solvable, if there exists an abelian tower in $G$ such that the last element is the trivial subgroup.

Proposition 7.5.5. The following propositions regard implications of the solvability of finite groups:
(a) Let $G$ be a finite group. An abelian tower of $G$ admits a cyclic refinement.
(b) Let $G$ be a finite solvable group. Then $G$ admits a cyclic tower whose last element is the trivial subgroup.

Proof. Let $G$ be a finite abelian group. We'll set out to prove that $G$ admits a cyclic tower ending with the trivial subgroup. We proceed by induction on the order of $G$. If $|G|=1$, then the proposition follows trivially. Suppose, as the hypothesis of induction, that the proposition is true for $1 \leqslant|G|<n$. Let now $|G|=n$ and consider any element $x \neq e$ of $G$. Define the group $X:=\langle x\rangle$, and $H:=G / X-$ since $|H|<|G|=n$ it must be true, from hypothesis, that there exists a cyclic tower

$$
H \supsetneq H_{0} \supsetneq \cdots \supsetneq H_{m}=\{e\} .
$$

Let $\pi: G \rightarrow H$ be the canonical projection. From Corollary 7.5.2, the sequence of subgroups $G_{0}:=G$, and $G_{j}:=f^{-1}\left(H_{j}\right)$ for $1 \leqslant j \leqslant m$ form a cyclic tower in $G$, thus

$$
G=G_{0} \supsetneq G_{1} \supsetneq \cdots \supsetneq G_{m} \supsetneq\{e\} .
$$

is the desired cyclic tower on $G$. This proves both the first statement and the second. 4
For the time being, the following theorem is a classic but probably still beyond the scope of this text, so I shall only mention it.

Theorem 7.5.6 (Feit-Thompson). All finite groups of odd order are solvable.
Theorem 7.5.7. Let $N \subseteq G$ be a normal subgroup of the group $G$. Then $G$ is solvable if and only if both $N$ and $G / N$ are solvable.

Proof. $(\Rightarrow)$ Suppose $G$ is solvable, and let $G=G_{0} \supsetneq \cdots \supsetneq G_{n}=\{e\}$ be an abelian tower in $G$. For each $1 \leqslant j \leqslant n$, define $N_{j}:=N \cap G_{j}$, and let $\phi_{j}: G_{j} \rightarrow Q$ be a group morphism with kernel $G_{j+1}$, then the induced morphism $\left.\phi_{j}\right|_{N \cap G_{j}}: N \cap G_{j} \rightarrow Q$ has kernel $N \cap G_{j+1}=N_{j+1}$ - which proves that $N_{j+1}$ is normal in $N_{j}$. Moreover, there exists a canonical embedding $N_{j} / N_{j+1} \hookrightarrow G_{j} / G_{j+1}$, which implies that $N_{j} / N_{j+1}$ is abelian by Proposition 7.3.12. Therefore $N=N_{0} \supsetneq \cdots \supsetneq N_{n}=\{e\}$ is an abelian tower and hence $N$ is solvable. For the group $G / N$ we can simply consider the tower $G / N=\left(G_{0} N\right) / N \supsetneq \cdots \supseteq\left(G_{n} N\right) / N=(\{e\} N) / N=N$ - where $\left(G_{j+1} N\right) / N$ is surely
normal in $\left(G_{j} N\right) / N$. Moreover, $\left[\left(G_{j} N\right) / N\right] /\left[\left(G_{j+1} N\right) / N\right] \simeq\left(G_{j} N\right) /\left(G_{j+1} N\right)$ which inherits the commutative structure from $G_{j} / G_{j+1}$ - thus the tower is abelian and $G / N$ is solvable.
$(\Leftarrow)$ Let both $N$ and $G / N$ be solvable and consider the abelian towers $G / N=H_{0} \supsetneq$ $\cdots \supsetneq H_{n}=N$, and $N=Q_{0} \supsetneq \cdots \supseteq Q_{m}=\{e\}$. Since every $H_{j}$ is isomorphic to a subgroup $H_{j}^{\prime}$ of $G$ containing $N$, we see that

$$
G=H_{0}^{\prime} \supsetneq \cdots \supsetneq H_{n}^{\prime}=N
$$

is an abelian tower. Moreover, we can append the abelian tower of $N$ to obtain the abelian tower

$$
G=H_{0}^{\prime} \supsetneq \cdots \supsetneq H_{n}^{\prime}=Q_{0} \supsetneq \cdots \supsetneq Q_{m}=\{e\} .
$$

This shows that $G$ is a solvable group.

## Commutator Group

Definition 7.5.8 (Commutator group). Let $G$ be a group, we define the commutator group $[G, G]$ of $G$ to be the subgroup of $G$ generated by all elements of the form $x y x^{-1} y^{-1}$, for $x, y \in G$.

Lemma 7.5.9. The commutator group $[G, G]$ is normal in $G$.
Proof. We show that $g[G, G] g^{-1}=[G, G]$ for all $g \in G$. Let $g, x, y \in G$ be any triple of elements. If we let $p:=g x g^{-1}$ and $q:=g y g^{-1}$ we obtain

$$
p q p^{-1} q^{-1}=\left(g x g^{-1}\right)\left(g y g^{-1}\right)\left(g x g^{-1}\right)\left(g y g^{-1}\right)=g x y x^{-1} y^{-1} g^{-1}
$$

therefore $g[G, G] g^{-1} \subseteq[G, G]$. Moreover, if $a:=g^{-1} x g$ and $b:=g^{-1} y g$, then

$$
\begin{align*}
g\left[a b a^{-1} b^{-1}\right] g^{-1} & =g\left[\left(g^{-1} x g\right)\left(g^{-1} y g\right)\left(g^{-1} x g\right)\left(g^{-1} x^{-1} g\right)\left(g^{-1} y^{-1} g\right)\right] g^{-1} \\
& =g\left[g^{-1} x y x^{-1} y^{-1} g\right] g^{-1} \\
& =x y x^{-1} y^{-1},
\end{align*}
$$

which implies in $x y x^{-1} y^{-1} \in g[G, G] g^{-1}$ and hence $[G, G] \subseteq g[G, G] g^{-1}$.
Lemma 7.5.10. The quotient group $G /[G, G]$ is commutative.
Proof. Notice that $x y=y x$ is equivalent to $x y(y x)^{-1}=x y x^{-1} y^{-1}=e$, thus, for any given $x, y \in G$, since $x y x^{-1} y^{-1} \in[G, G]$, then $[x y]=[y x] \in G /[G, G]-$ which shows the commutativity.

Lemma 7.5.11. Let $G$ be any group and $H$ be a commutative group. Any morphism of groups $f: G \rightarrow H$ has the commutator $[G, G]$ in its kernel. Therefore the following diagram commutes


Proof. If $H$ is commutative, then given any pair $x, y \in G$ we have

$$
f\left(x y x^{-1} y^{-1}\right)=f(x) f(y) f\left(x^{-1}\right) f\left(y^{-1}\right)=f(y) f(x) f(x)^{-1} f(y)^{-1}=e_{H}
$$

thus indeed $[G, G] \subseteq \operatorname{ker} f$.

### 7.6 Ab - Category of Abelian Groups

Definition 7.6.1 (Category of abelian groups). We define the category of abelian groups, denoted Ab , to be the category whose objects are abelian groups and group morphisms between them.

In the following propositions we are going to describe ways of telling if the group you might be interested is abelian or not.

Proposition 7.6.2. A group $G$ is abelian if and only either one of the following conditions are satisfied:

- The map $\phi: G \rightarrow G$ sending $g \mapsto g^{-1}$ is a morphism of groups.
- The map $\psi: G \rightarrow G$ sending $g \mapsto g^{2}$ is a morphism of groups.

Proof. If $G$ is abelian, then for any $g, g^{\prime} \in G$ we have:

$$
\begin{gathered}
\phi\left(g g^{\prime}\right)=\left(g g^{\prime}\right)^{-1}=g^{\prime-1} g^{-1}=g^{-1} g^{\prime-1}=\phi(g) \phi\left(g^{\prime}\right) \\
\psi\left(g g^{\prime}\right)=\left(g g^{\prime}\right)^{2}=\left(g g^{\prime}\right)\left(g g^{\prime}\right)=g^{2} g^{\prime 2}=\psi(g) \psi\left(g^{\prime}\right)
\end{gathered}
$$

That is, $\phi$ and $\psi$ are morphisms. Now suppose $\phi$ and $\psi$ are morphisms and consider any elements $g, g^{\prime} \in G$ - we get the following relations from each of the morphisms:

$$
\begin{gathered}
g g^{\prime}=\phi\left(g^{-1}\right) \phi\left(g^{\prime-1}\right)=\phi\left(g^{-1} g^{\prime-1}\right)=\phi\left(\left(g^{\prime} g\right)^{-1}\right)=g^{\prime} g \\
\left(g g^{\prime}\right)^{2}=\psi\left(g g^{\prime}\right)=\psi(g) \psi\left(g^{\prime}\right)=g^{2} g^{\prime 2}
\end{gathered}
$$

Using cancellation law for the second relation we find $g^{\prime} g=g g^{\prime}$. This shows that $G$ is abelian.

Proposition 7.6.3. Let $\phi \in \operatorname{Mor}_{G r p}\left(G, \operatorname{Aut}_{G r p}(G)\right)$ be the morphism of groups mapping $g \stackrel{\phi}{\mapsto} \gamma_{g}$ — as defined in Lemma 7.3.19. G is an abelian group if and only if $\phi$ is trivial. Proof. Let $G$ be an abelian group, then for all $g \in G$ the corresponding inner automorphism $\gamma_{g}$ maps $a \stackrel{\gamma_{g}}{\longmapsto} g a g^{-1}=a\left(g g^{-1}\right)=a$ hence $\gamma_{g}=\operatorname{id}_{G}$, thus $\phi$ is indeed trivial. Let $\phi$ be trivial, then for all $g \in G$ the map $\gamma_{g}=\operatorname{id}_{G}$ and therefore for all $a \in G$ we have $g a g^{-1}=a$, which implies in $g a g^{-1}=a g g^{-1}=a g^{-1} g=g g^{-1} a=g^{-1} g a$ — that is, the group $G$ is abelian.

Now that we know some ways of identifying abelian groups, we dive deep again into the categorical foundations of the category of abelian groups Ab.

Proposition 7.6.4 $\left(\operatorname{Mor}_{\mathrm{Ab}}\right.$ abelian group). Let $G, H \in \mathrm{Ab}$ be any two commutative groups. The collection of morphisms $\operatorname{Mor}_{\mathrm{Ab}}(G, H)$ forms an abelian group with a binary operation defined by $(\phi+\psi)(g)=\phi(g)+_{H} \psi(g)$ - where $+_{H}$ is the binary operation of $H$.

Proof. Let $\phi, \psi \in \operatorname{Mor}_{\mathrm{Ab}}(G, H)$ be any morphisms and consider elements $g, g^{\prime} \in G$. From the commutativity of $H$ it follows that

$$
\begin{aligned}
(\phi+\psi)\left(g+_{G} g^{\prime}\right) & =\phi\left(g+_{G} g^{\prime}\right)+_{H} \psi\left(g+_{G} g^{\prime}\right) \\
& =\left(\phi(g)+_{H} \phi\left(g^{\prime}\right)\right)+_{H}\left(\psi(g)+_{H} \psi\left(g^{\prime}\right)\right) \\
& =\left(\phi(g)+_{H} \psi(g)\right)+_{H}\left(\phi\left(g^{\prime}\right)+_{H} \psi\left(g^{\prime}\right)\right) \\
& =(\phi+\psi)(g)+_{H}(\phi+\psi)\left(g^{\prime}\right) .
\end{aligned}
$$

That is, $\phi+\psi \in \operatorname{Mor}_{\mathrm{Ab}}(G, H)$. Moreover, given a morphism $f \in \operatorname{Mor}_{\mathrm{Ab}}(G, H)$, define the map $k: G \rightarrow H$ mapping $k(g)=f(g)^{-1}$. Notice that $k$ is a morphism:

$$
\begin{aligned}
k\left(g+_{G} g^{\prime}\right) & =f\left(g+_{G} g^{\prime}\right)^{-1} \\
& =\left(f(g)+_{H} f\left(g^{\prime}\right)\right)^{-1} \\
& =f\left(g^{\prime}\right)^{-1}+_{H} f(g)^{-1} \\
& =f(g)^{-1}+_{H} f\left(g^{\prime}\right)^{-1} \\
& =k(g)+_{H} k\left(g^{\prime}\right) .
\end{aligned}
$$

Moreover, $(f+k)(g)=f(g)+_{H} k(g)=f(g)+_{H} f(g)^{-1}=e_{H}$ — that is, $f+k$ is the trivial morphism $g \mapsto e_{H}$, i.e. $k$ is the inverse of $f$. This finishes the proof that $\operatorname{Mor}_{\mathrm{Ab}}(G, H)$ is a group. For the commutativity, it follows directly from the commutativity of $H$ : for all $\phi, \psi \in \operatorname{Mor}_{\mathrm{Ab}}(G, H)$ we have

$$
(\phi+\psi)(g)=\phi(g)+_{H} \psi(g)=\psi(g)+_{H} \phi(g)=(\psi+\phi)(g) .
$$

The following corollaries follow immediately from the construction of the category.
Corollary 7.6.5. Let $G \in \operatorname{Grp}$ and $H \in \mathrm{Ab}$. The collection of morphisms $\operatorname{Mor}_{G r p}(G, H)$ forms a group under the binary operation defined above.

Corollary 7.6.6. Let $A \in \operatorname{Set}$ and $H \in \mathrm{Ab}$. The collection of morphisms $\operatorname{Mor}_{\text {Set }}(A, F H)$ forms a group under the binary operation defined above - where $F: \mathrm{Grp} \rightarrow$ Set is a forgetful functor.

Proposition 7.6.7 (Cokernel in Ab). The category of abelian groups has cokernels.
Proof. Let $\phi: G \rightarrow H$ be a morphism of abelian groups and $Q$ be any other abelian group and $\alpha: G \rightarrow Q$ be any morphism for which $\alpha \phi=0$ - the trivial morphism.

Since $G$ is abelian, it follows that any subgroup of $G$ is normal - in particular, im $\phi$ is normal in $G$ and from Corollary 7.4.18 we see that

thus we conclude that

$$
G / \operatorname{im} \phi \simeq \operatorname{coker} \phi
$$

A much more trivial proof of Proposition 7.4 .27 can be achieved easily in the category of abelian groups with the help of Proposition 7.6.7 - for expository purposes we shall not use our results concerning the equivalence of epimorphisms and surjection in the category of groups, as we proved above.

Proposition 7.6.8. Let $\phi: G \rightarrow H$ be a morphism of abelian groups. The following propositions are equivalent in Ab :
(a) The morphism $\phi$ is an epimorphism.
(b) The coker $\phi$ is trivial.
(c) The set-function $\phi$ is surjective.

Proof. - (a) $\Rightarrow(b)$ : Let $\phi$ be an epimorphism and consider both the canonical projection and the trivial morphism $\pi, 0: G \rightrightarrows$ coker $\phi$. Notice that, from the definition of the cokernel, both maps are trivial when composed with $\phi$, that is, $\pi \phi=0 \phi$ - but since $\phi$ is an epimorphism, then $\pi=0$ and such a thing can only occur when coker $\phi$ itself is trivial.

- (b) $\Rightarrow(\mathrm{c})$ : Let coker $\phi$ be trivial, that is, $G / \operatorname{im} \phi=\operatorname{im} \phi$, which implies that $G=\operatorname{im} \phi$ and hence $\phi$ is surjective.
- $(\mathrm{c}) \Rightarrow(\mathrm{a})$ : If $\phi$ is surjective, then since Ab is a concrete category, $\phi$ is an epimorphism.

Proposition 7.6.9 (Coequalizers in Ab ). The category of abelian groups have coequalizers. Moreover, given group morphisms $f, g: A \rightrightarrows B$ between abelian groups, we have

$$
\operatorname{coeq}(f, g)=B / \operatorname{im}(f-g)
$$

Proof. Given any abelian group $G$ and a map $m: B \rightarrow G$ such that $m f=m g$, we have that, for all $a \in A$,

$$
m((f-g)(a))=m(f(a)-g(a))=m f(a)-m g(a)=0
$$

therefore $\operatorname{im}(f-g) \subseteq \operatorname{ker} m$. Thus using the universal property Corollary 7.4.18 we obtain a unique morphism $n: B / \operatorname{im}(f-g) \rightarrow G$ such that the following diagram commutes


Therefore $B / \operatorname{im}(f-g)=\operatorname{coeq}(f, g)$.

## Coproduct and Fiber Product

Proposition 7.6.10 (Coproduct in Ab ). The direct product of abelian groups is a coproduct in Ab. That is, for any $G, H, W \in \operatorname{Ab}$ and morphisms $f \in \operatorname{Mor}_{\mathrm{Ab}}(W, G)$ and $k \in \operatorname{Mor}_{\mathrm{Ab}}(W, H)$, there exists a unique morphism $\varphi \in \operatorname{Mor}_{\mathrm{Ab}}(W, G \times H)$ such that the following diagram commutes


Where we define inclusion morphisms $g \stackrel{\iota_{G}}{\longmapsto}\left(g, e_{H}\right)$ and $h \stackrel{\iota_{H}}{\longmapsto}\left(e_{G}, h\right)$.
Proof. Since $\mathrm{Ab} \subset \mathrm{Grp}$ — that is, the category of abelian groups is a subcategory of Grp — then, from Proposition 7.3.4 there exists a functor $\mathrm{Ab} \rightarrow$ Set. Since coproducts exists in Set and are unique, a set-function $\varphi$ exists and is unique, commuting the diagram in Set. We now show that we can extend such set-function into a morphism of groups. Let $\varphi: G \times H \rightarrow W$ be the mapping $(g, h) \stackrel{\varphi}{\mapsto} f(g) k(h)$. Notice that

$$
\begin{aligned}
\varphi\left((g, h)\left(g^{\prime}, h^{\prime}\right)\right)=\varphi\left(g g^{\prime}, h h^{\prime}\right)=f\left(g g^{\prime}\right) k\left(h h^{\prime}\right)=f(g) f\left(g^{\prime}\right) k(h) k\left(h^{\prime}\right) & =f(g) k(h) f\left(g^{\prime}\right) k\left(h^{\prime}\right) \\
& =\varphi(g, h) \varphi\left(g^{\prime}, h^{\prime}\right)
\end{aligned}
$$

that is, $\varphi$ is a morphism of groups.
Remark 7.6.11 (Coproducts in Grp). Proposition 7.6.10 is not at all true for the category of groups. Consider for instance the cyclic groups $C_{2}=\left\{e_{x}, x\right\}$ and $C_{3}=\left\{e_{y}, y, y^{2}\right\}$. Let $\sigma_{k} \in S_{3}$, for $0 \leqslant k \leqslant 2$ be the rotation of $\{1,2,3\}$ by $k$, that is, the permutations represented by

$$
M_{\sigma_{0}}=\left[\begin{array}{lll}
1 & 0 & 0 \\
0 & 1 & 0 \\
0 & 0 & 1
\end{array}\right] \quad M_{\sigma_{1}}=\left[\begin{array}{lll}
0 & 0 & 1 \\
1 & 0 & 0 \\
0 & 1 & 0
\end{array}\right] \quad M_{\sigma_{2}}=\left[\begin{array}{lll}
0 & 1 & 0 \\
0 & 0 & 1 \\
1 & 0 & 0
\end{array}\right]
$$

Consider the embeddings $f: C_{2} \mapsto S_{3}$ mapping $x^{k} \stackrel{f}{\mapsto} \sigma_{k}$ for $k \in\{0,1\}$, and $g: C_{3} \mapsto S_{3}$ mapping $y^{k} \stackrel{g}{\mapsto} \sigma_{k}$ for $k \in\{0,1,2\}$.

Suppose, for the sake of contradiction, that $C_{2} \times C_{3}$ is a coproduct in Grp, that is, exists a unique morphism $\varphi: C_{2} \times C_{3} \rightarrow S_{3}$ such that $f=\varphi \iota_{C_{2}}$ and $g=\varphi \iota_{C_{3}}$. Since $\varphi$ is supposedly a morphism of groups,

$$
\begin{aligned}
\varphi(x, y) & =\varphi\left(x, e_{y}\right) \varphi\left(e_{x}, y\right)=\sigma_{1} \sigma_{1}=\sigma_{2} \\
\varphi\left(x, y^{2}\right) & =\varphi\left(x, e_{y}\right) \varphi\left(e_{x}, y^{2}\right)=\sigma_{1} \sigma_{2}=\sigma_{0}
\end{aligned}
$$

However, $\varphi\left(e_{x}, e_{y}\right)=\sigma_{0}$ and on the other hand we have $\varphi(x, y) \varphi\left(x, y^{2}\right)=\sigma_{2}$, which contradicts the properties of a group morphism. This shows that there exists no such $\varphi$ in Grp and hence $C_{2} \times C_{3}$ is not a coproduct in Grp.

Although $C_{2} \times C_{3}$ is not a coproduct in Grp, that doesn't mean that Grp has no coproducts, they just behave differently when compared with Ab. For instance, let $C_{2} * C_{3} \in \operatorname{Grp}$ be defined to be the group generated by elements $x$ and $y$, such that $x^{2}=e$ and $y^{3}=e$. We'll now show that $C_{2} * C_{3}$ is a coproduct of $C_{2}$ and $C_{3}$ in Grp. Let $G$ be any group and consider morphisms $f: C_{2} \rightarrow G$ and $k: C_{3} \rightarrow G$. The inclusions $\iota_{C_{2}}: C_{2} \rightarrow C_{2} * C_{3}$ and $\iota_{C_{3}}: C_{3} \rightarrow C_{2} * C_{3}$ will be naturally given maps by taking each element to itself.

Let $q \in C_{2} * C_{3}$ be any element. We know that there exists a finite collection of coefficients $I=\left\{(a, b) \in \mathbf{Z}^{2}\right\}$ such that $q=\prod_{(a, b) \in I} x^{a} y^{b}$. Define $\phi: C_{2} * C_{3} \rightarrow G$ as the mapping

$$
\phi(q)=\phi\left(\prod_{(a, b) \in I} x^{a} y^{b}\right)=\prod_{(a, b) \in I} f\left(x^{a}\right) k\left(y^{b}\right)=\prod_{(a, b) \in I} f(x)^{a} k(y)^{b}
$$

It should be clear that this definition implies $\phi \iota_{C_{2}}=f$ and $\phi \iota_{C_{3}}=g$. Notice that $\phi(e)=$ $f(e) k(e)=e_{G}$ and for all $q, p \in C_{2} * C_{3}$ - with respective coefficients $I=\left\{(a, b) \in \mathbf{Z}^{2}\right\}$ and $J=\left\{(c, d) \in \mathbf{Z}^{2}\right\}$ - we have

$$
\begin{aligned}
\phi(q) \phi(p)=\phi\left[\prod_{(a, b) \in I} x^{a} y^{b}\right] \phi\left[\prod_{(c, d) \in J} x^{c} y^{d}\right] & =\prod_{(a, b) \in I} f\left(x^{a}\right) k\left(y^{b}\right) \prod_{(c, d) \in J} f\left(x^{c}\right) k\left(y^{d}\right) \\
& =\prod_{(\alpha, \beta) \in A} f\left(x^{\alpha}\right) g\left(y^{\beta}\right) \\
& =\phi\left(\prod_{(\alpha, \beta) \in A} f\left(x^{\alpha}\right) g\left(y^{\beta}\right)\right) \\
& =\phi\left(\prod_{(a, b) \in I} x^{a} y^{b} \prod_{(c, d) \in J} x^{c} y^{d}\right)=\phi(q p)
\end{aligned}
$$

that is, $\phi$ is a morphism of groups - where we define $A$ as the concatenation of the coefficients $I$ and $J$. We have shown that the following diagram commutes


We can then conclude that $C_{2} * C_{3}$, as defined above, is the coproduct of $C_{2}$ and $C_{3}$ in Grp.

Proposition 7.6.12 (Fiber products). Fiber products exist in Ab. That is, given abelian groups $G, H, W \in \operatorname{Ab}$ and group morphisms $\phi \in \operatorname{Mor}_{\mathrm{Ab}}(G, W)$ and $\psi \in \operatorname{Mor}_{\mathrm{Ab}}(H, W)$. Let

$$
G \times_{W} H \in \mathrm{Ab}_{W}
$$

be the fiber product of $\phi$ and $\psi$ in the category $\mathrm{Ab}_{W} \subseteq \mathrm{Ab}^{4}$ for which exists natural projections $\pi_{G}$ and $\pi_{H}$. Let $Q \in \mathrm{Ab}$ be any abelian group and consider any morphisms $f \in \operatorname{Mor}_{\mathrm{Ab}}(Q, G)$ and $k \in \operatorname{Mor}_{\mathrm{Ab}}(Q, H)$. Then there exists a unique morphism $\varphi \in$ $\operatorname{Mor}_{\mathrm{Ab}}(Q, P)$ such that the following diagram commutes


Proof. Define $G \times_{W} H=\{(g, h) \in G \times H: \phi(g)=\psi(h)\}$. Since $\phi$ and $\psi$ are group morphisms, given $(g, h) \in G \times_{W} H$, we have

$$
\phi\left(g^{-1}\right)=\phi(g)^{-1}=\psi(h)^{-1}=\psi\left(h^{-1}\right),
$$

that is, $\left(g^{-1}, h^{-1}\right) \in G \times_{W} H$ exists and is the inverse of $(g, h)$. Moreover, given any elements $(g, h),\left(g^{\prime}, h^{\prime}\right) \in G \times_{W} H$, the product $(g, h)\left(g^{\prime}, h^{\prime}\right)=\left(g g^{\prime}, h h^{\prime}\right)$ is such that

$$
\phi\left(g g^{\prime}\right)=\phi(g) \phi\left(g^{\prime}\right)=\psi(h) \psi\left(h^{\prime}\right)=\psi\left(h h^{\prime}\right)
$$

thus $\left(g g^{\prime}, h h^{\prime}\right) \in G \times_{W} H$. This shows that $G \times_{W} H$ is indeed a group and since $G \times H$ is abelian, so is the fiber product defined above.

[^10]From the forgetful functor $\mathrm{Ab} \rightarrow$ Set we know that there exists a unique setfunction $\varphi$ such that the diagram commutes in Set. If we define $\varphi$ as the mapping $q \stackrel{\varphi}{\longmapsto}(f(q), k(q))$ we can see that it preserves the group structure, since

$$
\begin{aligned}
\varphi\left(q q^{\prime}\right) & =\left(f\left(q q^{\prime}\right), k\left(q q^{\prime}\right)\right) \\
& =\left(f(q) f\left(q^{\prime}\right), k(q) k\left(q^{\prime}\right)\right) \\
& =(f(q), k(q))\left(f\left(q^{\prime}\right), k\left(q^{\prime}\right)\right) \\
& =\varphi(q) \varphi\left(q^{\prime}\right)
\end{aligned}
$$

that is, $\varphi \in \operatorname{Mor}(\mathrm{Ab})$.

## Direct Sums and Free Groups

Definition 7.6.13 (Direct sum of abelian groups). Let $\left\{G_{j}\right\}_{j \in J}$ be an indexed collection of abelian groups. We define their direct sum as the collection of tuples $\left(g_{j}\right)_{j \in J}$ such that $g_{j} \neq e_{G_{j}}$ for only finitely many indexes $j \in J$ - that is, set of tuples with finite support. We denote the direct sum of $\left\{G_{j}\right\}_{j \in J}$ as $\bigoplus_{j \in J} G_{j}$ and the group structure of the direct sum is defined naturally as $\left(x_{j}\right)_{j \in J}+\left(y_{j}\right)_{j \in J}=\left(x_{j}+G_{j} y_{j}\right)_{j \in J}$.

Proposition 7.6.14 (Direct sum universal property). The direct sum defined in Definition 7.6 .13 satisfies the universal property of direct sums. In other words, let $\left\{G_{j}\right\}_{j \in J}$ be a collection of abelian groups and $H$ be any abelian group, in addition, consider the collection of morphisms $\left\{\phi_{j} \in \operatorname{Mor}_{G r p}\left(G_{j}, H\right)\right\}$. There exists a unique morphism $\psi: \bigoplus_{j \in J} G_{j} \rightarrow H$ such that the following diagram commutes

for every $j \in J$ - where $\iota_{j}: G_{j} \rightarrow \bigoplus_{j \in J} G_{j}$ is the natural inclusion, mapping $x_{j_{0}} \stackrel{\iota_{j}}{\mapsto}$ $\left(x_{j}\right)_{j \in J}$ such that $x_{j}=x_{j_{0}}$ for $j=j_{0}$ and $x_{j}=e_{G_{j}}$ for $j \neq j_{0}$.
Proof. Let $\psi: \bigoplus_{j \in J} G_{j} \rightarrow H$ be the map defined by $\psi(x)=\prod_{j \in J} \phi_{j}\left(x_{j}\right)$, where $x=$ $\left(x_{j}\right)_{j \in J} \in \bigoplus_{j \in J} G_{j}$ is any element. Notice that $\psi(x)$ is therefore a finite product of elements of $H$ and from the group structure of the direct sum, we find that $\psi$ is clearly a group morphism. Moreover, $\psi$ satisfies the commutativity $\psi \iota_{j}=\phi_{j}$ for each $j \in J$. Since such definition of $\psi$ defines the image of every element of its domain, $\psi$ is the unique morphism making the diagram commute.

If $G$ is an abelian group and $A, B \subseteq G$ are subgroups such that $A \cap B=0$ and $A+B=G$ - that is, every $g \in G$ can be written as $g=a+b$ for some $a \in A$ and $b \in B$ - then, in the context of abelian groups, we'll denote this fact shortly by $G=A \oplus B$.

Just like a vector space, we can define a basis of an abelian group by means of the ring $\mathbf{Z}$. Moreover, if an abelian group has a basis, then we say that it is free.

Definition 7.6.15 (Basis). Let $G$ be an abelian group. We say that a non-empty collection $\left\{g_{j}\right\}_{j \in J}$ is a basis for $G$ if, given an element $g \in G$, there exists a unique tuple of coefficients $\left(a_{j}\right)_{j \in J} \in \bigoplus_{j \in J} \mathbf{Z}$ such that $g=\sum_{j \in J} a_{j} g_{j}$.

Therefore the existence of a basis allow us to couple the coefficients coming from $\bigoplus_{j \in J} \mathbf{Z}$ to any element of $G$ in a unique way, which induces a natural isomorphism

$$
G \simeq \bigoplus_{j \in J} \mathbf{Z}
$$

Definition 7.6.16 (Abelian free group). An abelian group is said to be free if it allows a basis.

Equivalently as we did with free vector spaces, we can build free abelian groups out of sets, say $S$, by analysing maps $S \rightarrow \mathbf{Z}$ with finite support - the collection of those will be likewise denoted by $\mathbf{Z}^{\oplus S}$, which is a group under pointwise addition. For each $s \in S$ we define a map $\mathbf{s} \in \mathbf{Z}^{\oplus}$ by the mapping

$$
\mathbf{s}(x):= \begin{cases}1, & \text { for } x=s \\ 0, & \text { otherwise }\end{cases}
$$

Then, any element $\phi \in \mathbf{Z}^{\oplus S}$ can be written as a linear combination of finitely many maps $\mathbf{s}$ (which is possible because of the finite support of $\phi$ ) that is, for some $a_{j} \in \mathbf{Z}$ for each $1 \leqslant j \leqslant n$, we have

$$
\phi=\sum_{j=1}^{n} a_{j} \mathbf{s}_{j}, \text { mapping } s \stackrel{\phi}{\mapsto} \begin{cases}a_{j}, & \text { if } s=s_{j} \text { for some } 1 \leqslant j \leqslant n \\ 0, & \text { otherwise }\end{cases}
$$

Moreover, such choice of coefficients $a_{j} \in \mathbf{Z}$ is unique. Let $\left\{b_{j}\right\}_{j=1}^{n} \subseteq \mathbf{Z}$ be another set of coefficients with the same property, then in particular $\sum_{j=1}^{n}\left(a_{j}-b_{j}\right) \mathbf{s}_{j}=0$ and since $\mathbf{s}_{j}$ are all non-zero maps, we find that $b_{j}=a_{j}$. We also define a map $\iota_{S}: S \mapsto \mathbf{Z}^{\oplus S}$ by pairing $s \mapsto \mathbf{s}$.

Proposition 7.6.17 (Free Ab universal property). Let $S$ be a set. Consider $G$ to be any abelian group and a set-function $g: S \rightarrow G$, then there exists a unique morphism of groups $g_{*}: \mathbf{Z}^{\oplus S} \rightarrow G$ such that

is a commutative diagram.
Proof. Define $g_{*}$ by $g_{*}\left(\sum_{s \in S} a_{s} \mathbf{s}\right):=\sum_{s \in S} a_{s} g(s)$. Then $g_{*} \iota_{S}(s)=g_{*}(\mathbf{s})=g(s)$. Moreover, it follows from the construction that $g_{*}\left(\sum_{s \in S} a_{s} \mathbf{s}\right)=\sum_{s \in S} a_{s} g_{*}(\mathbf{s})$ thus $g_{*}$ is a group morphism. If we let $f: \mathbf{Z}^{\oplus S} \rightarrow G$ be a morphism satisfying such commutativity, we'll
see that, since $f_{*}(\mathbf{s})=g(s)=g_{*}(\mathbf{s})$, since any element of $\mathbf{Z}^{\oplus S}$ can be written uniquely as a linear combination of the basis $\{\mathbf{s}\}_{s \in S}$, then $f_{*}$ and $g_{*}$ agree in every point of the domain - hence $f_{*}=g_{*}$.

Corollary 7.6.18. Let $f: X \rightarrow Y$ be a set-function between sets $X$ and $Y$. Then, there exists a unique morphism $\bar{f}: \mathbf{Z}^{\oplus X} \rightarrow \mathbf{Z}^{\oplus Y}$ such that the following diagram commutes


Proof. Let $\bar{f}$ be defined by $\bar{f}\left(\sum a_{x} \mathbf{x}\right):=\sum a_{X} \iota_{Y} f(x)$ - so that, in particular, $\bar{f}(\mathbf{x})=\iota_{Y} f(x)$, that is, the diagram commutes. Let $\bar{g}: \mathbf{Z}^{\oplus X} \rightarrow \mathbf{Z}^{\oplus Y}$ be a morphism such that the diagram commutes, then necessarily $\bar{g}(\mathbf{x})=\iota_{Y} f(x)=\bar{f}(\mathbf{x})$, that is, $\bar{g}$ and $\bar{f}$ agree on the basis of $\mathbf{Z}^{\oplus X}$, thus $\bar{g}=\bar{f}$.

Notation 7.6.19. When it's not confusing, we can even drop the notation $\mathbf{s}$ and instead identify the elements as $s \in \mathbf{Z}^{\oplus S}$, so that $\sum a_{s} s:=\sum a_{s} \mathbf{s}$. Moreover, for now on, we'll refer to the free abelian group generated by $S$, that is, $\mathbf{Z}^{\oplus S}$, as $F_{\mathrm{Ab}}(S)$ - this is motivated by the fact that the notation $\mathbf{Z}^{\oplus S}$, although cool, may be a rather obscure way of talking about a group. The elements of $S$, in particular, will be referred to as free generators.

Proposition 7.6.20. Let $G$ be any abelian group
Continue: mini exercises on free ab grp

## Continue direct products

### 7.7 Graded Abelian Groups

Definition 7.7.1 (Graded abelian group). A graded abelian group is defined to be an abelian group $A$ consisting of an indexed collection of abelian groups $\left(A_{s}\right)_{s \in S}$-it should be noted that $S$ can be any set with an additive structure, for instance $A$ would be a Z-graded abelian group if $S=\mathbf{Z}$. An element $a \in A$ is said to have degree $s_{0}$ if $a \in A_{s_{0}}$.

Given $S$-graded abelian groups $A$ and $B$ we define a morphism $f: A \rightarrow B$ to be a collection of morphisms of groups $\left(f_{s}: A_{s} \rightarrow B_{s+d}\right)_{s \in S}$ for some fixed element $d \in S$, in such case we say that $f$ has degree $d$.

### 7.8 Free Groups

Let $S$ be a set. We define the category C as composed of objects that are set-functions $f: S \rightarrow G$, denoted $(f, G) \in C$, where $G \in \operatorname{Grp}$. Given another object $(g, H) \in C$,
a morphism between $f \rightarrow g$ is a morphisms of groups $\phi: G \rightarrow H$ such that the following diagram commutes


Proposition 7.8.1 (Free group universal property). Given a set $S$ we define the free group of $S$ to be the initial object $(\iota, F(S))$ in the category $C$ defined above. In other words, $F(S)$ is said to be the free group of $S$ if there exists a set-functions $\iota$ such that for all $(f, G) \in C$ there exists a unique morphism $f \rightarrow \iota$ given by a morphism of groups $\phi: F(S) \rightarrow G$ such that the diagram commutes


Although we defined the universal property of free groups, there is still no certainty that such objects shall in fact exist. We now proceed by proving that any set has a corresponding free group - which is unique up to isomorphism. To do that, we first prove the following lemma.

Lemma 7.8.2. Let $G$ be a group generated by the set-function $f: S \rightarrow G$, where $S$ is a set. Then there exists an indexing set $I$ and a collection of indexed groups $\left\{G_{i}\right\}_{i \in I}$ for which there exists $i \in I$ such that $G \simeq G_{i}$.

Proof. Since $G=\langle\operatorname{im} f\rangle$ then, if $S$ is a finite set, it follows $\operatorname{im} f$ is also finite, hence $G$ is either finite or enumerably infinite, since every element of $G$ is given by a finite product of elements in $\operatorname{im} f$ and their inverses. On the other hand, if $S$ is infinite, then $\operatorname{im} f$ can be either finite or infinite, on both of these cases we find that $G$ can be either finite or infinite - that is, in both of these cases we have $|G| \leqslant|S|$.

Define $X$ to be an infinitely enumerable set if $S$ is finite, on the other hand, if $S$ is infinite, let $|X|=|S|$. Let $A \subseteq X$ be non-empty and define $\Gamma(A)$ to be the collection of binary operations $\gamma: A \times A \rightarrow A$ such that $(A, \gamma)$ is a group. Define $\mathcal{A}=\{(A, \gamma): A \subseteq X, \gamma \in \Gamma(A)\}$, the collection of groups on subsets of the set $X$.

We now show that $\mathcal{A}$ satisfies the proposition. The set $X$ is defined so that every possible product of elements of $S$ can be injectively assigned to an element of $X$, that is, for every collection $B$ of finite sequences of elements of $S$, there exists an injective setfunction $i: B \mapsto X$. Since $\langle\operatorname{im} f\rangle$ can be seen set-wise as a collection of finite sequences of elements of $S$, then there exists a bijection $j:\langle\operatorname{im} f\rangle \stackrel{\cong}{\leftrightharpoons} B$ - for some $B$ as defined above. Since there exists an injection $i: B \rightarrow X$, the induced map $i: B \xrightarrow{\simeq} \operatorname{im} i \subseteq X$ is a bijection and hence the composition $i j:\langle\operatorname{im} f\rangle \stackrel{\simeq}{\Longrightarrow} \operatorname{im} i$ is also bijective. Then, given $\gamma \in \Gamma(\operatorname{im} i)$ the group $(\operatorname{im} i, \gamma)$ is isomorphic to $\langle\operatorname{im} f\rangle=G$.

Proposition 7.8.3 (Every set has a unique free group). Let $S$ be any set. Then there exists a free group $(\iota, F(S)$ ), unique up to isomorphism, such that $F=\langle\operatorname{im} \iota\rangle$ and $\iota$ is injective.

Proof. Let $I$ be an indexing set and $\left\{G_{i}\right\}_{i \in I}$ be an indexed collection of groups. Define

$$
F_{0}=\prod_{i \in I} \prod_{\ell \in \operatorname{Mor}}{ }_{s e t}\left(S, G_{i}\right)<
$$

and consider the set-function $\iota_{0}: S \rightarrow F_{0}$ defined by mapping the elements $s \in S$ to a tuple of pairs whose $j$-th component would be $\left(\ell_{j}(s), \ell_{j}\right) \in \prod_{\ell \in \operatorname{Morset}\left(S, G_{j}\right)} G_{j} \times\{\ell\}$.

Consider $G$ to be a group and let $g: S \rightarrow G$ generate $G$. From Lemma 7.8.2 we find that there exists a group $G_{j} \in\left\{G_{i}\right\}_{i \in I}$ such that there is an isomorphism $\phi: G \stackrel{ }{\cong} G_{j}$. Moreover, the set-function $\psi=\phi g: S \rightarrow G_{j}$ is an element of $\bigcup_{i \in I} \operatorname{Mor}_{\text {Set }}\left(S, G_{i}\right)$. Define the projection map $\pi_{j, \psi}: F_{0} \rightarrow G_{j} \times\{\psi\}$. We can now define a map $\psi_{*}=\phi^{-1} \pi_{j, \psi}: F_{0} \rightarrow G$ such that the following diagram commutes


Define the subgroup $F=\left\langle\operatorname{im} \iota_{0}\right\rangle$ of $F_{0}$ and let $\iota: S \rightarrow F$ be the set-function mapping $\iota(s)=\iota_{0}(s)$. Also, let $g_{*}: F \rightarrow G$ be the restriction $g_{*}=\left.\psi_{*}\right|_{F}=\left.\phi^{-1} \pi_{j, \psi}\right|_{F}$ - which immediately implies in the uniqueness of $g_{*}$. Summarizing, we may view this construction as the following diagram


We see that $(\iota, F)$ satisfies Proposition 7.8 .1 and thus is a free group of $S$. Moreover, since the construction works with no restriction on the choice of group $G$ and setfunction $g$, given any $s, s^{\prime} \in S$ distinct elements, define $G=\left\langle s, s^{\prime}\right\rangle$ and let $g: S \rightarrow G$ be any set-function such that $g(s)=s$ and $g\left(s^{\prime}\right)=s^{\prime}$. We see that for the diagram to commute, it is necessary that $l(s) \neq l\left(s^{\prime}\right)$, otherwise $g$ would not be equal to the composition $g_{*} \iota$. This shows that for any two distinct elements of $S$, their image under $\iota$ is also distinct - that is, $\iota$ is injective on $S$, which proves the last assertion of the proposition.

Corollary 7.8.4. Given a set $S$, its corresponding free group is unique up to isomorphism.

Proof. It suffices to see that if $F$ is a free group of $S$, then it is the initial object of the category $C$, we defined above. From this, we can use Proposition 1.2.28 to obtain the immediate conclusion that $F$ is unique up to isomorphism in $C$.

Notice since every set has a corresponding free group via the latter theorem, we can view the free group as a covariant functor $F$ : Set $\rightarrow$ Grp such that $F(X)$ is the free group of $X \in$ Set - which is unique up to isomorphism - and, given a set-function $\phi: X \rightarrow Y$ between sets $X$ and $Y$, we assign $F(f)=\phi_{*}: F(X) \rightarrow F(Y)$. This whole construction is such that the following diagram commutes


In the above diagram we see that $F(X)$ is the free group of $X$ generated by the image of set-function $\iota_{X}$, which induces an injection of the set-elements of $X$ into a group structure. These elements are called the free generators of $F(X)$. The same remark is true for $F(Y)$. Moreover, the functor $F$ preserves injectivity and surjectivity of the maps $\phi$, which is stated in the following lemma.

Lemma 7.8.5. If the map $\phi: X \rightarrow Y$ is a surjective (or injective) set-function, then $\phi_{*}: F(X) \rightarrow F(Y)$ is a surjective (or injective) morphism of groups.

Proof. From Eq. (7.10) we have $\phi_{*} \iota_{X}=\iota_{Y} \phi$, hence, given any element $\gamma \in \operatorname{im} \iota_{Y} \subseteq F(Y)$, there exists $x \in X$ for which $\tau_{Y} \phi(x)=\gamma$ - which is true only because $\phi$ is surjective - thus $\phi_{*} \iota_{X}(x)=\gamma$. Since $F(Y)=\left\langle\operatorname{im} \iota_{Y}\right\rangle$, it follows that, for any given element $\gamma \in F(Y)$, there exists a finite sequence of elements $\left(\gamma_{j}\right)_{j=1}^{n}$ with $\gamma_{j} \in \operatorname{im} \iota_{Y}$ for which $\gamma=\prod_{j=1}^{n} \gamma_{j}$. From the fact that $\gamma_{j} \in \operatorname{im} \iota_{Y}$, we are able to a corresponding sequence $\left(x_{j}\right)_{j=1}^{n}$ of elements $x_{j} \in X$ such that $\iota_{\gamma} \phi\left(x_{j}\right)=\gamma_{j}$ for each $1 \leqslant j \leqslant n$ - which is always possible since $\gamma_{j} \in \operatorname{im} \iota \gamma$. Then we obtain

$$
\phi_{*}\left(\prod_{j=1}^{n} \chi_{j}\right)=\prod_{j=1}^{n} \phi_{*}\left(\chi_{j}\right)=\prod_{j=1}^{n} \gamma_{j}=\gamma
$$

and therefore $\prod_{j=1}^{n} \chi_{j} \in \phi_{*}^{-1}(\gamma)$, which proves that $\phi_{*}$ is surjective.
Now we prove the second part of the lemma, let $\phi: X \mapsto Y$ be injective. Given distinct elements $a, b \in X$, since $\iota_{X}$ and $\iota_{Y}$ are injective, then $\iota_{Y} \phi(a) \neq \iota_{Y} \phi(b)$ and thus $\phi_{*} \iota_{X}(a) \neq \phi_{*} \iota_{X}(b)$ - that is, $\phi_{*}$ is injective on the set im $\iota_{X}$ and since $\phi_{*}$ is a morphism of groups and multiplication is preserved, it follows that $\phi_{*}$ is in fact injective in $\left\langle\operatorname{im} \iota_{X}\right\rangle=F(X)$. Thus $\phi_{*}$ is injective.

Proposition 7.8.6. If the cardinality of the sets $X$ and $Y$ are equal, the free groups $F(X)$ and $F(Y)$ are isomorphic.

Proof. Notice that if $|X|=|Y|$ then there exists a bijection $f: X \rightarrow Y$ between such sets. From Lemma 7.8 .5 we see that the induced morphism of groups $f_{*}: F(X) \rightarrow F(Y)$ is both injective and surjective - thus bijective. Since bijections are isomorphisms in Grp, it follows that $F(X) \simeq F(Y)$.

Proposition 7.8.7 (Canonical factorization). Every group is the image of a morphism on a free group. That is, given a group $G$, there exists a canonical surjective morphism of groups $F \rightarrow G$ such that $F$ is a free group.

Proof. Let $S=\{g \in G\}$ be a set and consider the identity set-function $f: S \rightarrow G$. The induced group morphism $f_{*}: F(S) \rightarrow G$ is simply a projection, establishing a surjection between $F(S)$ and $G$ and hence $f_{*}(F(S))=G$.
$\square$

Our construction of free groups got rather abstract and avoided any concrete identification of the elements of a free group - the intention being of creating the most general setting possible. In fact, our construction is equivalent to that of the standard notion regarding the free group $F(S)$ as composed of finite strings of elements of a set $S$ together with their inverses - forming the so called "words". This idea boils down to the concept of the elements of $S$ being generators of the free group $F(S)$, while the collection of elements of $F(S)$ itself are the relations between the generator elements which provide the structure of a group to $F(S)$. In that view, we define the following concept.

Definition 7.8.8 (Groups from generators and relations). Let $S$ be a set and $R \subseteq F(S)$ be any subset of the elements of the free group. Let $N$ be the smallest subgroup of $F(S)$ containing $R$. Then we define $F(S) / N$ to be the group determined by the generators $S$ and the relations $R$.

## Coproducts in Grp

We now get back to the idea of constructing coproducts, not only in Ab, but also in Grp. We exemplified, in the infamous example coming from Remark 7.6.11, that such construction doesn't come trivially without the requirement of commutativity. Now we are ready for the construction of coproducts in Grp, which will be supported by the ideas developed on free groups.

Proposition 7.8.9. The category of groups Grp has coproducts.
Proof. Let $J$ be an indexing set for which there exists a corresponding indexed collection of groups $\left\{G_{j}\right\}_{j \in J}$. Similarly, let $\left\{S_{j}\right\}_{j \in J}$ be a collection of sets for which $S_{j}=G_{j}$ if $G_{j}$ is an infinite group, otherwise, if $G_{j}$ happens to be finite, we let $S_{j}$ be any infinitely enumerable set (for instance, $\mathbf{N}$ ). We define now $S$ to be any set such that $|S|=\left|\coprod_{j \in J} S_{j}\right|$.

Define $\Gamma$ as the collection of all binary operation $\gamma: S \times S \rightarrow S$ for which the pair $(S, \gamma)$ is a group - which we'll shortly denote by $S_{\gamma}$. For each relation $\gamma \in \Gamma$, define the collection

$$
\Phi_{\gamma}=\left\{\phi=\left\{\varphi_{j} \in \operatorname{Mor}_{G r p}\left(G_{j}, S_{\gamma}\right): j \in J\right\}\right\}
$$

that is, the collection composed of elements $\phi$ which are themselves collections of group morphisms $G_{j} \rightarrow S_{\gamma}$ for each $j \in J$. For each $\phi \in \Phi_{\gamma}$, the product $S_{\gamma} \times\{\phi\}$ is a group with the operation defined as $\gamma^{*}((x, \phi),(y, \phi))=(\gamma(x, y), \phi)$.

Now we construct a group $F_{0}$ defined as

$$
F_{0}=\prod_{\gamma \in \Gamma} \prod_{\phi \in \Phi_{\gamma}} S_{\gamma} \times\{\phi\}
$$

which has a binary operation $F_{0} \times F_{0} \rightarrow F_{0}$ which is defined naturally as applying $\gamma^{*}$ to the corresponding $S_{\gamma} \times\{\phi\}$ factor of the tuples.

Notice that we carried over the collection $\phi$ to be able to relate $F_{0}$ to the groups $\left\{G_{j}\right\}_{j \in J}$ that we started with - to realize that, define the collection of morphisms

$$
I=\left\{\iota_{j} \in \operatorname{Mor}_{\operatorname{Grp}}\left(G_{j}, F_{0}\right): j \in J\right\}
$$

such that the mapping $\iota_{j}$ is defined by mapping elements $g \in G_{j}$ to the tuple of $F_{0}$ whose $S_{\gamma} \times\{\phi\}$ factor is given by $\left(\varphi_{j}(g), \phi\right)$ - recall that $\varphi_{j} \in \phi$ is a morphism of groups $\varphi_{j}: G_{j} \rightarrow S_{\gamma}$. So far, what we have is the following scenario


We set out to construct a group that would satisfy the coproduct universal property. We need to somehow modify both groups $F_{0}$ and $S_{\gamma}$ so that we have a unique arrow for any choice of a group $G$ in place of $S_{\gamma}$.

We start out by letting $G$ be any group and considering the collection

$$
\Psi=\left\{\psi_{j} \in \operatorname{Mor}_{G r p}\left(G_{j}, G\right): j \in J\right\}
$$

with the purpose of replacing both $S_{\gamma}$ and $\phi$, respectively. Let $H \subseteq G$ be the subgroup of $G$ generated by the images the morphisms of $\Psi$ - that is, $H=\left\langle\bigcup_{j \in J} \mathrm{im} \psi_{j}\right\rangle$. It is natural to see that since $H$ can be interpreted as the group whose elements are finite product of elements from $\left\{G_{j}\right\}_{j \in J}$ - and some of those can even be finite groups then $|H| \leqslant|S|$, since the latter was constructed so that $S$ had a cardinality bigger than or equal to the set-theoretic coproduct $\coprod_{j \in J} G_{j}$. From Lemma 7.8.2 we find that there exists $S_{\gamma} \in\left\{S_{\gamma}\right\}_{\gamma \in \Gamma}$ and an isomorphism of groups $\eta: S_{\gamma} \xrightarrow{\simeq} H$.

Let $i: H \hookrightarrow G$ be the canonical inclusion sending $H \ni g \mapsto g \in G$. Let $\phi \in \Phi_{\gamma}$ be the collection of morphisms such that $i \eta \varphi_{j}=\psi_{j}$, that is, the following diagram commutes for all $j \in J$


Define $p_{(\gamma, \phi)}: F_{0} \rightarrow S_{\gamma}$ to be a projection mapping the $(s, \phi) \in S_{\gamma} \times\{\phi\}$ factor into $s \in S_{\gamma}$. Define $\psi_{*}: F_{0} \rightarrow G$ to be the morphism of groups such that the following
diagram commutes


Notice that, by Section 7.8, we obtain - for all $g \in G_{j}$

$$
\operatorname{inp}_{(\gamma, \phi)} \iota_{j}(g)=\operatorname{i\eta p}_{(\gamma, \phi)}\left(\left(\ldots,\left(\varphi_{j}(g), \phi\right), \ldots\right)\right)=i \eta \varphi_{j}(g)=\psi_{j}(g)
$$

But, from Section 7.8, since $\operatorname{iqp}_{(\gamma, \phi)}=\psi_{*}$, then we conclude that $\psi_{*} \iota_{j}=\psi_{j}$, that is, the following diagram commutes


Let $F \subseteq F_{0}$ be the subgroup $F=\left\langle\bigcup_{j \in J} \operatorname{im} \iota_{j}\right\rangle$. Then, if we consider the restriction $\psi_{*}: F \rightarrow G$, the morphism $\psi_{*}$ is now uniquely defined by the collection of morphisms $\psi$, that is, the diagram

commutes for every $j \in J$ — hence $F$ is the coproduct in the category of groups. $\quad$. Exercise 7.12 Aluffi

### 7.9 Group Actions on Sets

Definition 7.9.1 (Group action). Let $G$ be a group and $A \in C$ be an object in some category C. An action of $G$ on the object $A$ is a group morphism

$$
G \longrightarrow \operatorname{Aut}_{C}(A),
$$

that is, the elements of $G$, which are automorphisms in the category $B G$, define automorphisms in $A$-this shows that group actions are functors $\mathrm{BG} \rightarrow \mathrm{C}$ (see Example 1.4.12 for more). Left actions are covariant functors, while right actions are contravariant. Naturally the action is said to be faithful if the functor is faithful, that is, if $G \mapsto \operatorname{Aut}_{\mathrm{C}}(A)$ is injective.

Definition 7.9.2. Let $G$ be a group acting on a set $X$. We define the following concepts concerning such action:
(a) The action is said to be free if the identity $e_{G}$ is the only element fixing any of the elements of $X$. In other words, $G$ acts freely on $X$ if given any $g \in G$ for which there exists $x \in X$ with $g \cdot x=x$, then $g=e$ is the identity element.
(b) An action is said to be effective if the only member of $G$ acting trivially is the identity $e_{G}$.

Remark 7.9.3. Mind you, free actions are effective, but not all effective actions are free.
Definition 7.9.4 (Conjugation). Let $G$ be a group. We define a conjugation to be a group action of $G$ on itself, $G \times G \rightarrow G$, mapping

$$
(g, h) \longmapsto g h g^{-1}
$$

Theorem 7.9.5 (Cayley). Every group acts faithfully on some set. That is, there exists a set $A$ and an injective action $G \mapsto \operatorname{Aut}_{\text {set }}(A)=S_{A}$-where $S_{A}$ is the symmetry group of $A$-making $G$ a subgroup of $S_{A}$.

Proof. Simply consider the action $G \rightarrow \operatorname{Aut}_{\text {Grp }}(G)$ mapping $g \mapsto f_{g}$, where $f_{g}(x)=g x$ is the left-multiplication by $g$. This action is certainly faithful, making $G$ isomorphic to a subgroup of $\operatorname{Aut}_{G r p}(G)$.

Definition 7.9.6 (Opposite group). Given a group $G$, we define its opposite group $G^{\text {op }}$ to be composed of the elements of $G$, and endowed with a contravariant action $G \times G \rightarrow G$ mapping $(g, h) \mapsto g \cdot h:=h g$.

Corollary 7.9.7. The following are properties relating a group $G$ with its opposite group:
(a) The set-function $\phi: G^{\mathrm{op}} \rightarrow G$ mapping $g \mapsto g$ is an isomorphism of groups $G^{\mathrm{op}} \simeq G$ if and only if $G$ is commmutative.
(b) There exists a natural isomorphism of groups $G \simeq G^{\text {op }}$ even when $G$ is noncommutative.

Proof. We first prove item (a). If the said map is an isomorphism, then for all $g, h \in G$ we have

$$
g h=\phi(g h)=\phi(h \cdot g)=\phi(h) \phi(g)=h g,
$$

thus $G$ is commutative. Conversely, if we assume that $G$ is commutative then the map, besides being bijective, is also a group morphism, since

$$
\phi(g \cdot h)=\phi(h g)=h g=g h=\phi(g) \phi(h)
$$

For item (b), define a map $\psi: G \rightarrow G^{\text {op }}$ to be given by $g \mapsto g^{-1}$, which is bijective since inverses are unique. The set-function $\psi$ is also a group morphism since

$$
\psi(g h)=(g h)^{-1}=h^{-1} g^{-1}=g^{-1} \cdot h^{-1}=\psi(g) \cdot \psi(h)
$$

Therefore $G \simeq G^{\text {op }}$ via $\psi$.

Proposition 7.9 .8 (Left to right \& back). Let $G$ be a group and $A$ be a set. Given any left-action $\sigma: G \rightarrow \operatorname{Aut}_{C}(A)$, we can turn $\sigma$ into a unique corresponding right-action $\sigma^{\mathrm{op}}: G \rightarrow \operatorname{Aut}_{C}(A)$ given by $\sigma^{\mathrm{op}}(g)(a)=\sigma\left(g^{-1}\right)(a)$ for any $g \in G$ and $a \in A$. The conversion of a right-action into a left-action is also unique.

Proof. Let $g, h \in G$ and $a \in A$ be any elements. The map $\sigma^{\text {op }}$ is indeed a right action:

$$
\sigma^{\mathrm{op}}(g h)(a)=\sigma\left((g h)^{-1}\right)(a)=\sigma\left(h^{-1} g^{-1}\right)(a)=\sigma\left(h^{-1}\right)\left(\sigma\left(g^{-1}\right)(a)\right)=\sigma^{\mathrm{op}}(h)\left(\sigma^{\mathrm{op}}(g)(a)\right)
$$

The uniqueness comes from the fact that inverses are unique.
Definition 7.9.9 (Transitive action). A group $G$ is said to act transitively on a set $A$ if for all pairs of elements $a, b \in A$, there exists a group element $g$ such that $g a=b$

Definition 7.9.10 (Orbit \& Stabilizer). Given a group action $G \rightarrow \operatorname{Aut}_{\text {set }}(A)$ on a set $A$, the orbit of an element $a \in A$ under the action of $G$ is defined to be the set

$$
\operatorname{Orb}_{G}(a):=\{g a: g \in G\} \subseteq A
$$

The stabilizer of $a$ under the action of $G$ is the subgroup

$$
\operatorname{Stab}_{G}(a):=\{g \in G: g a=a\} \subseteq G
$$

Let $\sigma$ be any action of $G$ on $A$. Stabilizers indeed define a subgroup of $G$, notice that if $g, h \in \operatorname{Stab}_{G}(a)$, then $\sigma(g h)(a)=\sigma(g)(\sigma(h)(a))=\sigma(g)(a)=a$ thus $g h \in \operatorname{Stab}_{G}(a)$. Moreover, since $\sigma$ is a group morphism, we have $\sigma\left(g^{-1}\right)(a)=\sigma(g)^{-1}(a)=a$ thus $g^{-1} \in \operatorname{Stab}_{G}(a)$.

Definition 7.9.11 (G-Set category). Given a group $G$, we define a category G-Set whose objects are pairs $(\sigma, A)$-where $A$ is a set, and $\sigma: G \rightarrow \operatorname{Aut}_{\text {set }}(A)$ is a group action on $A$-and morphisms $\phi:(\sigma, A) \rightarrow(\rho, B)$ are set-functions $\phi: A \rightarrow B$ such that the following diagram commutes


That is, $\phi(\sigma(g)(a))=\rho(g)(\phi(a))$ —or put even more simply as $\phi(g a)=g \phi(a)$ when there is no chance of confusion. These functions are called G-equivariant.

Proposition 7.9.12. Let $(\sigma, A) \in G$-Set, where $A$ is non-empty and $\sigma$ is a transitive left-action on $A$. Then there exists an isomorphism

$$
(\sigma, A) \simeq\left(\ell, G / \operatorname{Stab}_{\ell}(a)\right)
$$

in $G$-Set, where $\ell$ is the left-multiplication of $G$ on $G / \operatorname{Stab}_{\ell}(a)$, and $a \in A$ is any element.

Proof. For the sake of brevity, define $S_{a}:=\operatorname{Stab}_{\ell}(a)$. Let $a \in A$ be any element, and define a set-function $\phi: G / S_{a} \rightarrow A$ given by $\phi\left(g S_{a}\right):=g a$. This function is well defined since, for any $g S_{a}=g^{\prime} S_{a}$ we have $g^{\prime-1} g \in S_{a}$ thus $\left(g^{\prime-1} g\right) a=a$, which implies in $g a=g^{\prime} a$. To show that $\phi$ is bijective, we construct its inverse: define $\psi: A \rightarrow G / S_{a}$ by mapping $g a \mapsto g S_{a}$, which is well defined because if $g a=g^{\prime} a$ then $g^{\prime-1} g a=a$ thus $g^{\prime-1} g \in S_{a}$ so that $g S_{a}=g^{\prime} S_{a}$ by definition. One sees easily that $\phi$ and $\psi$ are inverses of each other. Finally, the map $\phi$ is also equivariant since

$$
\phi\left(g\left(g^{\prime} S_{a}\right)\right)=\phi\left(\left(g g^{\prime}\right) S_{a}\right)=\left(g g^{\prime}\right) a=g\left(g^{\prime} a\right)=g \phi\left(g^{\prime} S_{a}\right)
$$

Corollary 7.9.13. Let $G$ be a finite group and $A$ be a set. For any action of $G$ on a group $A$, the orbit $\operatorname{Orb}_{G}(a)$ of any element $a \in A$ is a finite subset of $A$. Moreover the we have that $\left|\operatorname{Orb}_{G}(a)\right|$ divides $|G|$.

Proof. Given a $G$-set $(\sigma, A)$ for any action $\sigma$, if $a \in A$ is any element, then the restriction

$$
\bar{\sigma}: G \longrightarrow \operatorname{Aut}_{\mathrm{set}}\left(\operatorname{Orb}_{G}(a)\right),
$$

is a transitive action by the definition of the orbit-where $\bar{\sigma}(g)\left(g^{\prime} a\right):=\sigma(g)(g a)$ for all $g \in G$ and $g^{\prime} a \in \operatorname{Orb}_{G}(a)$.

Therefore by Proposition 7.9.12 we have $\left(\bar{\sigma}, \operatorname{Orb}_{G}(a)\right) \simeq\left(\ell, G / \operatorname{Stab}_{G}(g a)\right)$ in $G$-Set, for any $g a \in \operatorname{Orb}_{G}(a)$-which means the existence of a $G$-equivariant bijection between the sets $\operatorname{Orb}_{G}(a)$ and $G / \operatorname{Stab}_{G}(g a)$. From Proposition 7.4.6 we have $\left|G / \operatorname{Stab}_{G}(g a)\right|=$ $|G| /\left|\operatorname{Stab}_{G}(g a)\right|$ and therefore

$$
\left|\operatorname{Orb}_{G}(a)\right| \cdot\left|\operatorname{Stab}_{G}(g a)\right|=|G| .
$$

Theorem 7.9.14. Let $G$ be a group acting on a set $A$. Consider elements $a \in A, g \in G$, and define $b:=g a$. It follows that

$$
\operatorname{Stab}_{G}(b)=g \operatorname{Stab}_{G}(a) g^{-1}
$$

Proof. Let $h \in \operatorname{Stab}_{G}(a)$ be any element, then

$$
\left(g h g^{-1}\right) b=(g h)\left(g^{-1} b\right)=(g h) a=g(h a)=g a=b
$$

therefore $g h g^{-1} \in \operatorname{Stab}_{G}(a)$. Now if $\ell \in \operatorname{Stab}_{G}(b)$, we have $\ell b=b$ but since $b=g a$ then $\ell(g a)=g a$ thus multiplying by $g^{-1}$ in both sides we obtain $\left(g^{-1} \ell g\right) a=a$. Therefore, $\operatorname{Stab}_{G}(b)=g \operatorname{Stab}_{G}(a) g^{-1}$.

### 7.10 Topological Groups

## Construction and Properties

Study topological groups on Dieck

## Actions on Topological Spaces

Definition 7.10.1 (Orbit space). Let $X$ be a topological space, and a topological group $G$ together with a left action on $X$. We define the orbit space of the $G$-space $X$ to be the set

$$
X / G:=X / x \sim \operatorname{Orb}_{G}(x)
$$

together with the quotient topology given by the canonical projection $X \rightarrow X / G$.
Definition 7.10.2 ( $G$-stable). Given a $G$-space $X$, a subset $A \subseteq X$ is said to be $G$-stable if for every pair $(g, a) \in G \times A$ we have $g a \in A$.

### 7.11 Group Objects

Definition 7.11.1 (Group object). Let C be a category with (finite) products and with a terminal object 1. A triple $(G, m, e, i)$ is said to be a group object of C if:

- $G$ is an object of $C$.
- $m: G \times G \rightarrow G$ is a morphism of $C$ satisfying the commutativity of

in C. This arrow defines the notion of multiplication of group members.
- $e: 1 \rightarrow G$ is a morphism of $C$ making the diagram

commute in C. The map $e$ defines the notion of a neutral member of the group object.
- $i: G \rightarrow G$ is a morphism of $C$ such that-if $\Delta:=\operatorname{id}_{G} \times \mathrm{id}_{G}$ is the diagonal morphism of $G$-the diagram

is commutative in $C$. The map $i$ defines the notion of inversion of members of the group.


## Chapter 8

## Rings \& Modules

### 8.1 Introduction

Definition 8.1.1 (Ring). A ring $R$ is a set endowed with an additive structure $+: R \times R \rightarrow$ $R$ such that $(R,+)$ is an abelian group, moreover, $R$ is also endowed with a second multiplicative structure $: R \times R \rightarrow R$, making $(R, \cdot)$ into a monoid - that is, it is both associative and has a two-sided identity:

- Given $r, s, t \in R$, we have $r(s t)=(r s) t$.
- There exists a unitary element $1_{R} \in R$ for which all $r \in R$ is such that

$$
r 1_{R}=1_{R} r=r .
$$

Furthermore, we impose that the product structure is distributive over the additive structure, that is, given any three $r, t, s \in R$, one has

$$
(r+t) s=r s+t s \text { and } r(t+s)=r t+r s .
$$

Remark 8.1.2. Notice that we impose the following in the definition of a ring - which may vary from author to author - the ring is always associative and has a unity.

Definition 8.1.3. A ring $R$ is said to be commutative if for all $r, s \in R$ we have $r s=s r$.
Corollary 8.1.4. Given a ring $R$, its identity is unique.
Proof. Let $1,1^{\prime} \in R$ be both identities of $R$, then $11^{\prime}=1$ but $11^{\prime}=1^{\prime}$, thus $1=1^{\prime}$.
Corollary 8.1.5. Given a ring $R$, we have for all $r \in R$ that $0_{R} r=r 0_{R}=0_{R}$.
Proof. Note that for both cases, since $0=0+0$, we are able to write the following

$$
\begin{aligned}
& r \cdot 0=r \cdot(0+0)=r \cdot 0+r \cdot 0, \\
& 0 \cdot r=(0+0) \cdot r=0 \cdot r+0 \cdot r .
\end{aligned}
$$

Further, since $(R,+)$ is a group, we evoke the cancellation to obtain $r \cdot 0=0$ and $0 \cdot r=0$.

Corollary 8.1.6. Let $R$, then $(-1) \cdot r=r \cdot(-1)=-r$.
Proof. Notice that since $1 \cdot r=r \cdot 1=r$ we may write

$$
\begin{aligned}
& r+(-1) \cdot r=(1-1) \cdot r=0 \cdot r=0, \\
& r+r \cdot(-1)=r \cdot(1-1)=r \cdot 0=0 .
\end{aligned}
$$

Thus we arrive at the desired conclusion.

## Examples of Rings

Example 8.1.7 (Trivial ring). The trivial group * can be endowed with a ring structure by imposing that $* * *=*$ and $*+*=*$, so that $(*,+, \cdot)$ is a ring - we call such ring trivial or zero-ring. Moreover, in such ring we have the equality $0=1$.

Corollary 8.1.8. A ring $R$ is a zero-ring if and only if $0=1$.
Proof. If $R$ is a zero-ring, it's obvious that $0=1$. On the other hand, assume that $R$ is a ring such that $0=1$, then if $r \in R$ is any element, we have $r=1 \cdot r=0 \cdot r=0$, thus $R$ is indeed a zero ring.

Example 8.1.9 (Power set ring). Let $S$ be any set and $2^{S}$ be its power set. If we define additive and multiplicative structures in $2^{S}$ given by $A+B:=(A \cup B) \backslash(A \cap B)$ and $A \cdot B:=A \cap B$, for all $A, B \in 2^{S}$, then $\left(2^{S},+, \cdot\right)$ is a ring.

In order to prove that, we first show that $\left(2^{S},+\right)$ forms an abelian group. Since $A \cup B=B \cup A$ and $A \cap B=B \cap A$, then clearly $A+B=B+A$. Moreover, the empty set $\varnothing$ is the additive identity, since $A \cup \varnothing=A$ and $A \cap \varnothing=\varnothing$ and $A \backslash \varnothing=A$ - thus $A+\varnothing=A$. Every set is its own inverse: notice that $A \cup A=A$ and $A \cup A=A$, hence $A+A=A \backslash A=\varnothing$. We therefore conclude that $\left(2^{S},+\right)$ is an abelian group.

Now we show that $\left(2^{S}, \cdot\right)$ is a monoid with distributivity over addition. Let $A, B, C \in$ $2^{S}$ be any three sets. For associativity we have

$$
(A \cdot B) \cdot C=(A \cap B) \cap C=A \cap(B \cap C)=A \cdot(B \cdot C)
$$

The unitary element is $S$ itself, since $A \cap S=S \cap A=A$ for all $A \in 2^{S}$. Moreover,

$$
\begin{aligned}
(A+B) \cdot C & =[(A \cup B) \backslash(A \cap B)] \cap C \\
& =[(A \cap C) \cup(B \cap C)] \backslash[(A \cap C) \cap(B \cap C)] \\
& =A \cdot C+B \cdot C .
\end{aligned}
$$

We can analogously show the same for $A \cdot(B+C)=A \cdot B+A \cdot C$.
Example 8.1.10. Let $R$ be a ring and $S$ be a set. The set of set-functions $R^{S}$, of the form $S \rightarrow R$, together with the pointwise addition and multiplication make $R^{S}$ into a ring.

Example 8.1.11. Let $M_{n}(R)$ denote the collection of all $n \times n$ matrices with entries in the ring $R$. If we endow $M_{n}(R)$ with component-wise addition and matrix multiplication, then $\left(M_{n}(R),+, \cdot\right)$ is a ring.

## Zero-Divisors

Definition 8.1.12 (Zero-divisor). Let $R$ be a ring. We say that $a \in R$ is a left-zero-divisor if there exists $b \neq 0$ in $R$ for which $a b=0$. On the other hand, $a$ is said to be a right-zero-divisor if $b a=0$.

Remark 8.1.13. The zero-ring is the only ring with no zero-divisors.
If $R$ is a non-zero ring, then $0 \in R$ and since 0 is clearly a zero-divisor, $R$ has at least one zero-divisor. On the other hand, if $*$ is a zero-ring, then zero is the only element of * - thus we have no element different than zero for there to be a zero-divisor.

Proposition 8.1.14. Let $R$ be a ring. An element $a \in R$ is not a left-zero-divisor (or right-zero-divisor) if and only if the map $R \rightarrow R$ given by the left-multiplication by $a$ (or right-multiplication) is injective.

Proof. We only concern ourselves with the left case, the right case is clearly analogous. Let $a \in R$ be a non-left-zero-divisor, then if $a b=a c$ for some $b, c \in R$, then $a b-a c=$ $a(b-c)=0$ - and, since $a$ is not a zero-divisor, then necessarily $b-c=0$, that is $b=c$.

On the other hand, if $a$ is a left-zero-divisor then let $b \neq 0$ be any element of $R$ such that $a b=0$. Notice then that both $b$ and zero have the same image under the multiplication map, thus $R \rightarrow R$ is definitely not injective.

Remark 8.1.15. Notice that Proposition 8.1.14 simply states that for an element to be a zero-divisor the cancellation law must be true, that is, $a \in R$ is not a zero-divisor if $a b=a c$ implies $b=c$ for every $b, c \in R$.

Definition 8.1.16 (Domain). An object $R$ is said to be a domain if $R$ is non-zero ring and for every two elements $a, b \in R$ such that $a b=0$ we must have either $a=0$ or $b=0$.

Definition 8.1.17 (Integral domain). An object $R$ is said to be an integral domain if $R$ is a non-zero commutative ring such that, for all $a, b \in R$ such that $a b=0$, then either $a=0$ or $b=0$.

That is, in an integral domain every non-zero element is a non-zero-divisor. Classic examples of integral domains are $\mathbf{Z}, \mathbf{Q}, \mathbf{R}$ and $\mathbf{C}$.

Corollary 8.1.18. In an integral domain $R$, if $r, a, b \in R$ are such that $r a=r b$ and $r \neq 0$, then $a=b$ - that is, cancellation by non-zero elements holds.

Proof. This is immediate from Proposition 8.1.14.
Corollary 8.1.19. If $R$ is an integral domain, an element $x \in R$ is such that $x^{2}=1$, if and only if $x \in\{-1,1\}$.

Proof. If $x \in\{-1,1\}$, clearly $x^{2}=1$, since this is true for any given ring. For the other implication, suppose now that $x^{2}=1$, then $x^{2}-1=(x+1)(x-1)=0$ but since $R$ is an integral domain, either $x=1$ or $x=-1$.

## Nilpotent Elements

Definition 8.1.20 (Nilpotent). Let $R$ be a ring. An element $a \in R$ is said to be nilpotent if there exists some $n \in \mathbf{Z}$ such that $a^{n}=0$.

Lemma 8.1.21. Let $R$ be a ring. If $a, b \in R$ are nilpotent elements with $a b=b a$, then $a+b$ is nilpotent.

Proof. Suppose $n, m \in \mathbf{Z}$ are such that $a^{n}=b^{m}=0$. Since $a$ and $b$ commute, we can write the binomial equation $(a+b)^{n m}=\sum_{j=0}^{n m}\binom{n m}{j} a^{n m-j} b^{j}$ - for $j<m$ we have $n m-j>n m-m=n(m-1)$, greater than a multiple of $n$, so that $a^{n m-j}=0$; for $j \geqslant m$, we have $b^{j}=0$. Thus we arrive at $(a+b)^{n m}=0$.

Lemma 8.1.22. A class $[m]_{n}$ is nilpotent in $\mathbf{Z} / n \mathbf{Z}$ if and only if $m$ shares all prime factors of $n$.

Proof. Suppose that there exists a prime $p$ in the prime factorization of $n$ such that $p$ does not divide $m$. If $\ell \in \mathbf{Z}$ is any exponent, $m^{\ell}$ won't be divisible by $n$ since it lacks the $p$ factor - hence $m$ fails to be nilpotent.

If $m$ shares every prime factor of $n$, let $k$ be defined to be the maximum exponent in the prime factorization of $n$ (which is ensured to exist since the collection is finite). Then clearly $[m]_{n}^{k}=\left[m^{k}\right]_{n}=0$ since $m^{k}$ is divisible by $n$.

Lemma 8.1.23. Let $R$ be a commutative ring and $x \in R$ be a nilpotent element. The following statements hold:
(a) The element $x$ is either zero or zero-divisor.
(b) For every $r \in R$, the product $r x$ is nilpotent.
(c) For every invertible element $u \in R$, the sum $x+u$ is invertible.

Proof. (a) If $x$ is non-zero, both the left and right multiplications of $x$ have the same image under the distinct elements $x^{n-1}$ and $0-$ hence $x$ is a zero-divisor.
(b) From commutativity $(r x)^{n}=r^{n} x^{n}=r^{n} \cdot 0=0$.
(c) First we deal with the case where $u=1$. Remember that given any $a \in R$ and $m \in \mathbf{N}$ we have

$$
(1+a)\left(1-a+a^{2}-\cdots+(-1)^{m} a^{m}\right)=1+(-1)^{m} a^{m+1} .
$$

Therefore, for every $m>n$, we have $(1+x) \sum_{j=0}^{m}(-1)^{j} a^{j}=1$ - which shows that $1+x$ is invertible. For the general case, if $v$ is the inverse of $u$, one sees that for any $m>n$ we have

$$
(u+x) \sum_{j=0}^{m}(-1)^{j} v x^{j}=1+(-1)^{m} x^{m+1}=1 .
$$

## Units

Definition 8.1.24 (Unit). Let $R$ be a ring. An element $u \in R$ is said to be a left-unit (or right-unit) if there exists $a \in R$ for which $u a=1$ (or $a u=1$ ). An element is said to be a unit if it is both a left and right unit.

Proposition 8.1.25. Let $R$ be a ring and $u \in R$ be an element. The following propositions hold.
(a) The element $u$ is a left-unit (or right-unit) if and only if the left-multiplication (or right-multiplication) map $R \rightarrow R$ is surjective.
(b) If $u$ is a left-unit (or right-unit), then $u$ is a non-right-zero-divisor (or non-left-zerodivisor).
(c) The inverse of a unit is unique.
(d) Units form a group under multiplication - such group is denoted by $R^{\times}$.

Proof. (a) Suppose first that $u$ is a left-unit. Let $a \in R$ be any element and suppose that $v \in R$ is such that $u v=R$ (which is ensured to exist) - then $u(v a)=(u v) a=a$, thus left-multiplication is surjective.
On the other hand, if the left-multiplication by $u$ is surjective, we can conclude that there must exist $v \in R$ for which $u v=1$ - hence $u$ is a left-unit of $R$.
The proof for the right-unit case is completely analogous and won't be included.
(b) If $u$ is a left-unit, assume $v \in R$ is such that $u v=1$. Notice that the rightmultiplication maps $f_{u}, f_{v}: R \rightrightarrows R$ by $u$ and $v$, respectively, are such that - for all $a \in R$,

$$
f_{v} f_{u}(a)=f_{v}(a u)=(a u) v=a(u v)=a \cdot 1=a .
$$

Therefore, $f_{v}$ is the left-inverse of $f_{u}$ and thus $f_{u}$ is injective. The equivalent proof can be written for the right-unit.
(c) Let $u$ be a unit and $v_{1}, v_{2} \in R$ be such that $v_{1} u=1$ and $u v_{2}=1$ - notice that

$$
v_{1}=v_{1} \cdot 1=v_{1}\left(u v_{2}\right)=\left(v_{1} u\right) v_{2}=1 \cdot v_{2}=v_{2}
$$

(d) Let $U$ denote the collection of units of $R$ (such set is ensured to be non-empty since 1 is a unit) - zero-rings are not a problem since zero is a unit in such case. If $a, b \in U$ are units, then there are $x, y \in U$ such that $a x=x a=1$ and $b y=y b=1$ (from the last item, inverses are unique) - therefore, $(a b)(y x)=a(b y) x=a x=1$, that is, $a b \in U$ and $y x$ its inverse.

Lemma 8.1.26. Let $R$ be a ring and $a \in R$ be an element. If $a$ is a right-unit (or left-unit) and has two or more left-inverses (or right-inverses), then

- The element $a$ is not a left-zero-divisor (or right-zero-divisor).
- The element $a$ is a right-zero-divisor (or left-zero-divisor).

Proof. We shall restrict ourselves with the right-unit case - the other is completely analogous. Since $a$ is a right-unit, there must exist $u \in R$ such that $u a=1$, then if $x, y \in R$ are such that $a x=a y$, we can multiply both left-sides by $u$ and obtain (ua) $x=x=y=(u a) y$ - thus left-multiplication by $a$ is injective and $a$ is not a left-zero-divisor.

For the other item, notice that if $u$ and $v$ are distinct left-inverses of $a$, then $u a=$ $v a=1$ have the same image under right-multiplication by $a$ - hence $a$ is a right-zerodivisor.

## Division Rings \& Fields

Definition 8.1.27 (Division ring). A ring $R$ is said to be a division ring if every non-zero element is a unit.

Definition 8.1.28 (Field). A non-zero ring $R$ is said to be a field if $R$ is a commutative division ring.

Remark 8.1.29. By the item (b) of Proposition 8.1.25, we see that a unit is a non-zerodivisor, thus, if every non-zero element is a unit, then every non-zero element is also a non-zero-divisor. This implies directly that every field is an integral domain. On the other hand, $\mathbf{Z}$ is an integral domain but fails to be a field, hence not every integral domain is a field. The concepts can however coincide in a special case, as we see in Proposition 8.1.31.

Proposition 8.1.30. Any subring of a field is an integral domain.
Proof. Let $k$ be a field and $R \subseteq k$ be a non-zero subring. Certainly $R$ is commutative. Moreover if $a, b \in R$ are such that $a b=0$, suppose, for the sake of contradiction, that $a$ and $b$ are both non-zero. From the field properties, there are inverses $a^{-1}, b^{-1} \in k$ so that $a^{-1}(a b) b^{-1}=\left(a^{-1} a\right)\left(b b^{-1}\right)=1 \cdot 1=1$, but $a b=0$ - this implies that $1=0$, which is a contradiction, thus $a$ or $b$ are zero.

Proposition 8.1.31. A finite and commutative ring $R$ is a field if and only if $R$ is an integral domain.

Proof. If $R$ is a finite commutative integral domain, let $a \in R$ be a non-zero-divisor of $R$ - so that multiplication by $a$ is an injective map. Since the ring is finite, the multiplication by $a$ is also surjective - thus by Proposition 8.1.25 we find that $a$ is a unit.

Proposition 8.1.32 (Particular case of Weddenburn's theorem). Let $R$ be a division ring with $p^{2}$ elements and $p$ be a prime, then $R$ is a field.

Proof. For the proof, we shall use the concepts of centre and centralizer, developed in Definition 8.2.15 and Definition 8.2.18. If we assume that $R$ is non-commutative, the centre $Z(R)$ will be a proper subring of $R$. Moreover, since $Z(R)$ is also a subgroup of $R$ under multiplication, by Corollary 7.4 .7 we conclude that $|Z(R)|=p$. Let now $r \in R \backslash Z(R)$, clearly the centralizer must contain both the centre of the ring - since by
definition it must commute with $r$ - and by symmetry, $r \in Z(r)$, thus $\{r\} \cup Z(R) \subseteq Z(r)$. Since the centralizer is a subgroup of $R$ and $|Z(r)|>p$, one concludes by Corollary 7.4.7 that $|Z(r)|=p^{2}$ - therefore $Z(r)=R$. That, however gives us a contradiction because if the centralizer of $r$ is the whole ring, then $r$ commutes with every element and thus $r \in Z(R)$ - this cannot be the case by hypothesis, thus one concludes that $R$ must be commutative and therefore a field.

## Monoid Ring

Definition 8.1.33 (Monoid ring). Given a monoid ( $M, \cdot$ ) and a ring $R$, we define a new ring called monoid ring $R[M]:=R^{\oplus M}$ - that is, the ring whose elements are formal linear combinations $\sum_{m \in M} a_{m} \cdot m$, where the coefficients $a_{m} \in R$ are non-zero for at most finitely many $m \in M$.

The unitary element of $R[M]$ is defined to be $1_{R} \cdot 1_{M}$. The additive and multiplicative structures are defined as follows

$$
\begin{aligned}
\sum_{m \in M} a_{m} \cdot m+\sum_{m \in M} b_{m} \cdot m & :=\sum_{m \in M}\left(a_{m}+b_{m}\right) \cdot m \\
\sum_{m \in M} a_{m} \cdot m+\sum_{m \in M} b_{m} \cdot m & :=\sum_{m \in M} \sum_{m_{1} \cdot m_{2}=m}\left(a_{m_{1}} b_{m_{2}}\right) \cdot m .
\end{aligned}
$$

## Polynomial Ring

Definition 8.1.34 (Polynomial). Let $R$ be a ring. A polynomial is a map $f: R \rightarrow R$ given by a finite linear combination of non-negative powers of the indeterminate variable with coefficients in $R$

$$
f(x)=\sum_{j \geqslant 0} a_{j} x^{j}, \text { where } a_{j}=0 \text { for } j \gg 0 .
$$

The collection of all polynomials over $R$ with variable $x$ is denoted by $R[x]$ - this collection forms a ring under point-wise addition and product, that is, given another element $g(x):=\sum_{j \geqslant 0} b_{j} x^{j}$, we define

$$
\begin{aligned}
f(x)+g(x) & :=\sum_{j \geqslant 0}\left(a_{j}+b_{j}\right) x^{j}, \\
f(x) \cdot g(x) & :=\sum_{k \geqslant 0} \sum_{i+j=k} a_{i} b_{j} x^{i+j} .
\end{aligned}
$$

It should be noted immediately that $R[x] \simeq R[\mathbf{N}]$, where $\mathbf{N}$ is viewed as a monoid under addition.

Two polynomials are said to be equal if the sequence of coefficients match - that is, if $f(x):=\sum_{j \geqslant 0} a_{j} x^{j}$ and $g(x):=\sum_{j \geqslant 0} b_{j} x^{j}$, then $f=g$ if and only if $a_{j}=b_{j}$ for all $j \geqslant 0$ - since the collection of non-zero coefficients is necessarily finite, this equality relation is well defined.

Corollary 8.1.35. If $R$ is a commutative ring, then $R[x]$ is commutative.
Lemma 8.1.36. A ring $R$ is an integral domain if and only if $R[x]$ is an integral domain.
Proof. If $R$ is an integral domain, then clearly $R[x]$ is both non-zero and commutative. Moreover, suppose that $f(x):=\sum_{j \geqslant 0} a_{j} x^{j}$ and $g(x):=\sum_{j \geqslant 0} b_{j} x^{j}$ are elements of $R[x]$ such that $f \cdot g=0$ - notice that from definition we have

$$
f(x) g(x)=\sum_{k \geqslant 0} \sum_{i+j=k} a_{i} b_{j} x^{i+j},
$$

which is merely a finite combination with coefficients, so that if $f \cdot g=0$, we necessarily have $a_{j}=0$ for all $j \geqslant 0$ or $b_{j}=0$ for all $j \geqslant 0$ since $R$ is an integral domain - thus either $f=0$ or $g=0$.

On the other hand, if $R[x]$ is an integral domain, let $a, b \in R$ be elements such that $a b=0$ in $R$, then we can consider the constant polynomials $f(x):=a$ and $g(x):=b$ to conclude that, since $f \cdot g=0$ then $a=0$ or $b=0$.

Definition 8.1.37. The degree of a non-zero polynomial $f(x):=\sum_{j \geqslant 0} a_{j} x^{j}$ is defined as the maximum index for which the corresponding coefficient is non-zero, that is

$$
\operatorname{deg} f:=\max \left\{j: a_{j} \neq 0\right\}
$$

This is well defined because $a_{j} \neq 0$ for only finitely many indices. By convention, the zero-polynomial has degree $-\infty$.

Polynomials of multiple variables are obtained simply as an iteration of the process of construction of $R[x]$, that is,

$$
R\left[x_{1}, \ldots, x_{n}\right]:=R\left[x_{1}\right] \ldots\left[x_{n}\right] .
$$

The ordering of the list $x_{1}, \ldots, x_{n}$ is irrelevant in the construction of the ring, if we swap any elements of the list we still end up with isomorphic rings.

One can consider a ring of polynomials over infinitely many variables, say $R\left[x_{1}, x_{2}, \ldots\right]$, but still every polynomial of such ring consists of only finitely many terms - as we impose that the coefficient indexing set is finite.

Definition 8.1.38 (Power series ring). Let $R$ be a ring. The ring of power series with variable $x$ and coefficients in $R$, denoted by $R \llbracket x \rrbracket$, is defined to be the ring whose elements are formal sums $\sum_{j=0}^{\infty} a_{j} x^{j}$.

### 8.2 The Category of Rings

Definition 8.2.1. Let $R$ and $S$ be rings. We define a morphism of rings $\phi: R \rightarrow S$ to be a map such that, for all $a, b \in R$, we have $\phi(a+b)=\phi(a)+\phi(b)$ and $\phi(a b)=\phi(a) \phi(b)$. Moreover, we also impose that morphisms are unitary, that is, $\phi\left(1_{R}\right)=1_{S}$.

Definition 8.2.2. The category of rings is defined to consist of rings and morphisms between them - such category will be denoted by Ring.

Remark 8.2.3. Curiously, the zero-ring is not a zero-object in the category of rings, it is a final object but fails to be initial - this is due to the fact that we imposed morphisms to be unitary, hence for a non-zero ring $R$, there exists no ring morphism from the zero-ring to $R$.

One should, however, not be afraid, because Ring does have initial objects. In fact, $\mathbf{Z}$ is initial in Ring - for any ring $R$ we can define a unique ring morphism $\mathbf{Z} \rightarrow R$ mapping $n \mapsto n \cdot 1_{R}$ for any $n \in \mathbf{Z}$.

Corollary 8.2.4 (Unit preservation). If $\phi: R \rightarrow S$ is a ring morphism and $u \in R$ is a left-unit (or right-unit), then $\phi(u) \in S$ is again a left-unit (or right-unit).

Proof. Let $v \in R$ be such that $v u=1_{R}$, then $\phi(v u)=\phi(v) \phi(u)$ but $\phi(v u)=\phi\left(1_{R}\right)=1_{S}$ thus $\phi(v) \phi(u)=1_{S}$ and $\phi(u)$ is indeed a left-unit in $S$.

Proposition 8.2.5 (Image of morphism is a subring). The image of a ring morphism $R \rightarrow S$ is a subring of $S$.

Proof. Let $\phi: R \rightarrow S$ be any ring morphism. Since it is a ring morphism, $\phi\left(1_{R}\right)=1_{S} \in$ $\operatorname{im} \phi$, moreover, if $a, b \in R$ then $\phi(a+b)=\phi(a)+\phi(b) \in \operatorname{im} \phi$ and $\phi(a b)=\phi(a) \phi(b) \in$ $\operatorname{im} \phi$ - thus $\operatorname{im} \phi$ is a subring of $S$.

Remark 8.2.6. The image of non-zero-divisors may be a zero-divisor in the new ring for instance, $3 \in Z$ is a non-zero-divisor but the ring morphism $Z \rightarrow \mathbf{Z} / 6 \mathbf{Z}$ (canonical projection) maps $3 \mapsto[3]_{6}$, which is a zero divisor since $[2]_{6}[3]_{6}=[0]_{6}$.

Proposition 8.2.7. Let $\phi: R \rightarrow S$ be a set-function between rings $R$ and $S$ such that the additive and multiplicative structures are preserved. If either one of the following propositions hold, then $\phi$ is a morphism of rings:
(a) The set-function $\phi$ is surjective.
(b) The set-function $\phi \neq 0$ and $S$ is an integral domain.

Proof. We simply need to show that each of the propositions yield $\phi\left(1_{R}\right)=1_{S}$ — which is the only condition left for $\phi$ to be a morphism of rings.
(a) If $\phi$ is surjective, then there exists $r \in R$ such that $\phi(a)=1_{S}$ but then

$$
1_{S}=\phi(a)=\phi\left(a \cdot 1_{R}\right)=\phi(a) \phi\left(1_{R}\right)=1_{S} \cdot \phi\left(1_{R}\right)=\phi\left(1_{R}\right)
$$

(b) If $\phi\left(1_{R}\right)$ where to be zero, the whole image of $\phi$ would also evaluate to zero, thus $1_{R}$ has a non-zero image under $\phi$. Let $r \in R$ be any element such that $\phi(r) \neq 0$ - then $\phi(r)=\phi\left(r \cdot 1_{R}\right)=\phi(r) \phi\left(1_{R}\right)$ thus by means of Corollary 8.1.18 we can conclude that $\phi\left(1_{R}\right)=1_{S}$.

## Universal Property of Polynomial Rings

Let $A:=\left\{a_{1}, \ldots, a_{n}\right\}$ be a set and define a category $\mathrm{R}_{A}$ consisting of pairs $(\alpha, R)-$ where $\alpha: A \rightarrow \mathbf{R}$ is a set-function and $R$ is a ring such that $\alpha\left(a_{i}\right)$ commutes with every element of $R$ for each $1 \leqslant i \leqslant n$ (one may restrict this further by imposing that $R$ is commutative). Intuitively, morphisms between two objects $(\alpha, R) \rightarrow(\beta, S)$ are ring morphisms $\phi: R \rightarrow S$ such that the following diagram commutes in the category of sets:


In the same context of what is explained above, we now state a universal property concerning the ring $\mathbf{Z}\left[x_{1}, \ldots, x_{n}\right]$ in the category $\mathrm{R}_{A}$.

Proposition 8.2.8. Let $R$ be a ring and define a set-function $\alpha: A \rightarrow R$ mapping $a_{j} \mapsto x_{j}$. The object $\left(\alpha, \mathbf{Z}\left[x_{1}, \ldots, x_{n}\right]\right)$ is initial in $\mathrm{R}_{A}$.

Proof. Let $(\beta, R)$ be any object in $\mathrm{R}_{A}$. We now construct a morphism

$$
\phi:\left(\alpha, \mathbf{Z}\left[x_{1}, \ldots, x_{n}\right]\right) \rightarrow(\beta, R)
$$

We can impose that $\phi\left(x_{j}\right):=\beta\left(a_{j}\right)$ for every $1 \leqslant j \leqslant n$. For $\phi$ to be a ring morphism, one must also impose that is preserves both the additive and multiplicative structures of the rings. Since $\mathbf{Z}$ is initial in Ring, we can uniquely determine that $\phi(r)=\psi(r)$ where $\psi: \mathbf{Z} \rightarrow \mathbf{R}$ is a unique morphism. That is, we defined morphism $\phi$ given by

$$
\begin{aligned}
\phi\left(\sum m_{i_{1} \ldots i_{n}} x_{1}^{i_{1}} \ldots x_{n}^{i_{n}}\right) & =\sum \phi\left(m_{i_{1} \ldots i_{n}}\right) \phi\left(x_{1}^{i_{1}}\right) \ldots \phi\left(x_{n}^{i_{n}}\right) \\
& =\sum \psi\left(m_{i_{1} \ldots i_{n}}\right) \beta\left(a_{1}\right)^{i_{1}} \ldots \beta\left(a_{n}\right)^{i_{n}}
\end{aligned}
$$

which is surely both a ring morphism and unique - thus $\left(\alpha, \mathbf{Z}\left[x_{1}, \ldots, x_{n}\right]\right)$ is initial, and


Proposition 8.2.9 (Universal property of polynomial rings). Let $f: R \rightarrow S$ be any ring morphism and $s_{0} \in S$ be a fixed element commuting with $f(r)$ for all $r \in R$. There exists a unique ring morphism $\phi: R[x] \rightarrow S$ which extends $f$ and maps $x \mapsto s_{0}$.

Proof. In the construction of $\phi$ we first impose that $\phi(x):=s_{0}$. Moreover, for $\phi$ to extend $f$ one has to define $\phi(r):=f(r)$ for all given $r \in R$. In order for $\phi$ to be a ring
morphism it also needs to preserve the additive and multiplicative structures of the rings - hence we end up with

$$
\phi\left(\sum_{j \geqslant 0} r_{j} x^{j}\right)=\sum_{j \geqslant 0} \phi\left(r_{j}\right) \phi(x)^{j}=\sum_{j \geqslant 0} f\left(r_{j}\right) s_{0}^{j} .
$$

Since $\phi$ is completely defined by the image under constant polynomials and $x$, we find that $\phi$ is the unique ring morphism sending $x$ to $s_{0}$ for which the following diagram commutes


In fact, in the case where $R$ is commutative and $f: R \rightarrow R$ is an endomorphism, the unique ring morphism $\phi: R[x] \rightarrow R$ defined above is called the evaluation map at the given fixed point. We normally denote such map by eval ${ }_{r}$ - where $r \in R$ is the chosen fixed point for evaluation.

## Monomorphisms

Definition 8.2.10 (Kernel). The kernel of a ring morphism $\phi: R \rightarrow S$ is the set

$$
\operatorname{ker} \phi:=\left\{r \in R: \phi(r)=0_{S}\right\}
$$

Proposition 8.2.11 (Ring monomorphisms). Let $\phi: R \rightarrow S$ be a ring morphism. The following properties are equivalent:
(a) The ring morphism $\phi$ is a monomorphism.
(b) The kernel of $\phi$ is trivial - that is, $\operatorname{ker} \phi=\left\{0_{R}\right\}$.
(c) The set-function $\phi$ injective.

Proof. (a) $\Rightarrow$ (b): Suppose $\phi$ is monic. Let $r \in \operatorname{ker} \phi$ be any element and consider the uniquely defined morphisms eval,$_{r}$ eval $_{0_{R}}: \mathbf{Z}[x] \rightrightarrows R$ provided by Proposition 8.2.8 by fixing $\operatorname{eval}_{r}(x)=r$ and $\operatorname{eval}_{0_{R}}(x)=0_{S}$. Notice that $\phi \operatorname{eval}_{r}(n)=\phi \operatorname{eval}_{0_{R}}(n)=\phi(n)$ for any $n \in \mathbf{Z}$, while $\phi \operatorname{eval}_{r}(x)=\phi(r)=0_{S}$ and $\phi \operatorname{eval}_{0_{R}}(x)=\phi\left(0_{R}\right)=0_{S}$ - thus eval $_{r}=\operatorname{eval}_{0_{S}}$, which in turn implies that $r=\operatorname{eval}_{r}(x)=\operatorname{eval}_{0_{R}}(x)=0_{S}$. Therefore indeed $\operatorname{ker} \phi=\left\{0_{R}\right\}$.
$(\mathrm{b}) \Rightarrow(\mathrm{c})$ : Suppose $\phi$ has trivial kernel and let $r, t \in R$ be such that $\phi(r)=\phi(t)$, then $\phi(r)-\phi(t)=\phi(r-t)=0$ and $r-t \in \operatorname{ker} \phi$, which implies in $r=t$ - that is, $\phi$ is injective.
(c) $\Rightarrow$ (a): Suppose $\phi$ is an injective set-function. Since injections are monomorphisms in Set, we conclude that the set-function $\phi: R \rightarrow S$ is a monomorphism in Set when $R$ and $S$ are viewed as sets. If we now endows $R$ and $S$ with their respective ring structures, we obtain that $\phi$ is a monic in Ring.

## Subrings

Definition 8.2.12 (Subring). Let $R$ be a ring. A subring $S$ of $R$ is a ring whose elements are contained in $R$ and the canonical inclusion $\operatorname{map} S \hookrightarrow R$ is a ring morphism.

Remark 8.2.13. It should be noted that $S$ is a ring with unity where $1_{S}=1_{R}$.
Proposition 8.2.14 (Intersection of subrings). Let $R$ be a ring and $\left\{S_{j}\right\}_{j \in J}$ be a collection of subrings of $R$. The intersection $\bigcap_{j \in J} S_{j}$ is also a subring of $R$.
Proof. Since each $S_{j}$ is a subring, if $a, b \in \bigcap_{j \in J} S_{j}$ then $a, b \in S_{j}$ for all $j \in J$, and $a+b, a b \in S_{j}$. Since every subring contains the unity, so does the intersection - hence $\bigcap_{j \in J} S_{j}$ is a subring of $R$.

## Center

Definition 8.2.15 (Center). Given a ring $R$, we define its centre to be

$$
Z(R):=\{r \in R: r x=x r \text { for all } x \in R\} .
$$

Corollary 8.2.16. The centre of a ring is a subring.
Proof. Let $R$ be a ring. Clearly, $1 \in Z(R)$. Moreover, if $r, s \in Z(R)$ then for any $x \in R$ we have

$$
\begin{align*}
(r+s) x=r x+s x & =x r+x s=x(r+s) \\
(r s) x=r(s x)=(s x) r & =(x s) r=x(s r)=x(r s)
\end{align*}
$$

Therefore $Z(R)$ is indeed a subring of $R$.
Corollary 8.2.17. If $R$ is a division ring, then its centre $Z(R)$ is a field.
Proof. Since $Z(R)$ inherits the division ring structure and every element of $Z(R)$ commutes, it is indeed a field.

Definition 8.2.18 (Centralizer). Given a ring $R$ and an element $r \in R$, the centralizer of $r$ is defined to be the collection of elements $x \in R$ such that $x r=r x$ - we shall denote the centralizer of $r$ as $Z(r)$.

Corollary 8.2.19. The centralizer of an element is a subring.
Proof. This is simply a straightforward particular case of Corollary 8.2.16.
Corollary 8.2.20. In a division ring a centralizer is also a division ring.
Proof. Let $R$ be a division ring and $r \in R$ any element. If $x \in Z(r)$ is any element, then since $r x=x r$, we have $\left(r^{-1} x^{-1}\right) r x=\left(r^{-1} x^{-1}\right) x r=1$ thus $r^{-1} x^{-1} r$ is a left-inverse of $r$ - moreover, we equivalently see that $r x^{-1} r^{-1}$ is a right-inverse of $r$. Since $Z(r)$ is a subring, such inverses are contained in $Z(r)$ and thus $Z(r)$ is a division ring.

Notice however that the centralizer may not be a field since there can be noncommuting elements in $Z(r)$.

Corollary 8.2.21. The centre of a ring is the intersection of all centralizers of the ring.
Proof. Let $R$ be a ring. Certainly, if $x \in Z(R)$ then it commutes with every element of $R$ - which is equivalent to $x \in \bigcap_{r \in R} Z(r)$. Moreover, if an element belongs to the intersection of all centralizers, every element of the ring commutes with it and thus such element is also present in the centre of the ring. Therefore $Z(R)=\bigcap_{r \in R} Z(r)$. $\quad$,

## Epimorphisms

Remark 8.2.22 (Epimorphisms and surjection in Ring). In Ring, epimorphisms are not necessarily surjective set-functions.

A classical counterexample is the ring morphism given by the canonical inclusion $\iota: \mathbf{Z} \hookrightarrow \mathbf{Q}$. Let $R$ be any ring and consider ring morphisms $f, g: \mathbf{Q} \rightrightarrows R$ such that the following diagram commutes

$$
\mathbf{Z} \xrightarrow{\iota} \mathbf{Q} \xrightarrow[g]{\stackrel{f}{\rightrightarrows}} R
$$

Since $\left.f\right|_{\mathbf{Z}}=\left.g\right|_{\mathbf{Z}}$ we have, for any $p / q \in \mathbf{Q}$, that $f(p / q)=f(p) f(q)^{-1}$ and $g(p / q)=$ $g(p) g(q)^{-1}$, which implies in $f(p / q)=g(p / q)$ - that is, $f=g$ in general, which implies that $\iota$ is an epimorphism in Ring. For the shock of the reader, the same is obviously not the case in Set. If we take the rings $\mathbf{Z}$ and $\mathbf{Q}$ as abelian groups, one sees that coker $\phi$ is non-trivial and by Proposition 7.6 .8 we arrive at the fact that $\phi$ is not an epimorphism in Ab neither.

Remark 8.2.23. With the caution given by Remark 8.2.22 one can rightly observe that in the category of rings a morphism may be both monic and epic but yet lack the conditions for being an isomorphism.

Remark 8.2.24. Cokernels in Ring are not what one would normally expect of a good category, notice that given a ring morphism $\phi: R \rightarrow S$ and, if $\alpha: S \rightarrow Q$ is any ring morphism such that $\alpha \phi=0$ - as is required by the universal property of cokernels — then $\alpha \phi(1)=\alpha(1)=0$, which can only be the case for $Q=0$, the zero-ring. We therefore conclude that coker $\phi$ must be the zero-ring.

## Products

Proposition 8.2.25 (Product). Products exist in the category of rings.
Proof. Let $R$ and $S$ be rings. We'll define on $R \times S$ additive and multiplicative structures naturally as follows:

$$
\begin{gathered}
(x, a)+_{R \times S}(y, b):=\left(x+_{R} y, a+s b\right) \\
(x, a) \cdot R \times S(y, b):=(x \cdot R y, a \cdot s b)
\end{gathered}
$$

for any $x, y \in R$ and $a, b \in S$. We now check that $R \times S$ is a product in Ring.

Let $Z$ be any ring and let $\phi: Z \rightarrow R$ and $\psi: Z \rightarrow S$ be any two ring morphisms. Considering the canonical projections $\pi_{R}: R \times S \rightarrow R$ and $\pi_{S}: R \times S \rightarrow S$, we can construct a map $\ell: Z \rightarrow R \times S$ sending $z \mapsto(\phi(z), \psi(z))$. Such a map inherits the preservation of both the multiplicative and additive structures of $Z$ and $R \times S$ since $\phi$ and $\psi$ do so - therefore $\ell$ is a morphism of rings. Moreover, $\ell$ is completely determined by the image of both $\phi$ and $\psi$, hence $\ell$ is the unique morphism of rings such that the following diagram commutes


Remark 8.2.26. Coproducts on the other hand, although present in Ring, are not as easy to construct. Lets work out a special case: consider the commutative ring $\mathbf{Z}[x, y]$ in the category of commutative rings. We'll show that $\mathbf{Z}[x, y]$ is the coproduct of two copies of $\mathbf{Z}[x]$. Let $R$ be any commutative ring together with morphisms of rings $f, g: \mathbf{Z}[x] \rightrightarrows R$. We define a map $\phi: \mathbf{Z}[x, y] \rightarrow R$ for which $x \mapsto f(x)$, while $y \mapsto g(x)$ and finally $n \mapsto n$ for every $n \in \mathbf{Z}$. One immediately sees that $\phi$ indeed satisfies every condition of being a ring morphism. Moreover, since the image of $x, y$ and $\mathbf{Z}$ completely defines $\phi$, we conclude that $\phi$ is the unique morphism of rings such that the following diagram commutes

thus $\mathbf{Z}[x, y]$ is a coproduct of two copies of $\mathbf{Z}[x]$ in the category of commutative rings.

## The Ring $\operatorname{End}_{\mathrm{Ab}}(G)$

The ring of endomorphisms of an abelian group $\operatorname{End}_{\mathrm{Ab}}(G)$ is a structure that pops up in plenty of situations - module theory will explore this ring structure vastly so in this subsection we take some time to consider a small collection of interesting and useful facts about it. Such ring, has its additive structure defined by point-wise addition and the multiplicative structure is given by composition of morphisms.

Proposition 8.2.27. In the category of rings, there exists a natural isomorphism

$$
\mathbf{Z} \simeq \operatorname{End}_{\mathrm{Ab}}(\mathbf{Z})
$$

Proof. Let $\phi: \operatorname{End}_{\mathrm{Ab}}(\mathbf{Z}) \rightarrow \mathbf{Z}$ sending $f \mapsto f(1)$ it is clear that

$$
\phi(f+g)=(f+g)(1)=f(1)+g(1)=\phi(f)+\phi(g)
$$

moreover the multiplicative structure is also preserved:

$$
\phi(f g)=(f g)(1)=f(g(1))=f(1 \cdot g(1))=f(1) g(1)=\phi(f) \phi(g)
$$

Also, $\phi(\mathrm{id})=\mathrm{id}(1)=1$, therefore $\phi$ is a ring morphism. Since the image of the identity element over a group morphism completely determines the map, one can be certain that $\phi$ is a bijection - thus an isomorphism of rings.

Lets agree for the time being that, given a ring $R$ and an element $r \in R$, the morphisms of rings ${ }_{r} m, m_{r}: R \rightrightarrows R$ are the left and right, respectively, multiplication of elements of $R$ by $r$.

Proposition 8.2.28. Let $R$ be a ring. The map $m: R \mapsto \operatorname{End}_{\mathrm{Ab}}(R)$ sending $r \mapsto{ }_{r} m$ is an injective ring morphism.

Proof. We first show that $m$ is a ring morphism: let $a, b \in R$ be any three elements, then for any $r \in R$ we have

$$
m(a+b)(r)={ }_{a+b} m(r)=(a+b) r=a r+b r=m(a)(r)+m(b)(r)
$$

On the other hand, multiplication yields composition, as expected

$$
m(a b)(r)={ }_{a b} m(r)=(a b) r=a(b r)={ }_{a} m\left({ }_{b} m(r)\right)=m(a)(m(b)(r)) .
$$

Also, $m(1)(r)=1 \cdot r=r=\operatorname{id}_{R}(r)$ and hence $m(1)=\operatorname{id}_{R}$. We conclude that $m$ is indeed a ring morphism. The injectivity comes from the fact that the only element that yields a zero-map in $\operatorname{End}_{\text {Ring }}(R)$ is zero - thus the kernel is trivial.

Remark 8.2.29. Notice that one cannot further extend Proposition 8.2.28 for rightmultiplications, this comes from the fact that

$$
m_{a b}(r)=r(a b)=(r a) b=m_{b}\left(m_{a}(r)\right) .
$$

That is, the multiplicative structure has its order reversed when transitioning to the compositional structure.

Proposition 8.2.30. Up to isomorphism, there exists a unique ring (with identity) structure whose underlying group is $(\mathbf{Z},+)$.

Proof. Let $R$ be any ring with underlying group $\mathbf{Z}$, and fix any $r \in R$. Consider the ring morphism $m: R \rightarrow \operatorname{End}_{\mathrm{Ab}}(R)$ as defined in Proposition 8.2.28. Let $f \in \operatorname{End}_{\mathrm{Ab}}(R)$ is completely defined by $f\left(1_{R}\right)$, thus $f\left(1_{r}\right)^{m}=f$. Therefore $m$ is surjective and hence a bijection - $m$ thus establishes an isomorphism $R \simeq \operatorname{End}_{\mathrm{Ab}}(R)$. Moreover, by Proposition 8.2.27 we have $\operatorname{End}_{\mathrm{Ab}}(R) \simeq \mathbf{Z}$, thus

$$
R \simeq \operatorname{End}_{\mathrm{Ab}}(R) \simeq \mathbf{Z},
$$

which proves the statement.

Proposition 8.2.31. Let $R$ be a ring and. There is a subring $S \subseteq Z(R)$ of the centre of $R$ such that there exists a ring isomorphism

$$
S \simeq Z\left(\operatorname{End}_{\mathrm{Ab}}(R)\right)
$$

Proof. Let $f \in Z\left(\operatorname{End}_{\mathrm{Ab}}(R)\right)$ be any group morphism. Since $f$ commutes with every group endomorphism on $R$, in particular $f$ commutes with every right-multiplication by an element of $R$ - that is, given $r \in R$, we have $f(x r)=f(x) r$, which can only be the case if $f$ was a left-multiplication by an element of $R$. Moreover, by Proposition 8.2.28 the map $m: R \rightarrow \operatorname{End}_{\mathrm{Ab}}(R)$ is injective - restricting $m$ to the subring of $Z(R)$ given by $S:=\left\{r \in Z(R):{ }_{r} m \in Z\left(\operatorname{End}_{\mathrm{Ab}}(R)\right)\right\}$ makes $\left.m\right|_{S}$ into a surjective morphism. Therefore the morphism $\bar{m}: S \simeq Z\left(\operatorname{End}_{\mathrm{Ab}}(R)\right)$, where $\bar{m}(r)=m(r)$ for all $r \in S$, is an isomorphism of rings.

Corollary 8.2.32. Let $n \in \mathbf{Z}_{>0}$ be a positive integer. There exists a ring isomorphism

$$
\mathbf{Z} / n \mathbf{Z} \simeq \operatorname{End}_{\mathrm{Ab}}(\mathbf{Z} / n \mathbf{Z})
$$

### 8.3 Ideals \& Quotients of Rings

## Ideals

Definition 8.3.1 (Ideal). Let $R$ be a ring. A subgroup a of $(R,+)$ is said to be left-ideal of $R$ if for all $r \in R$ we have $r \mathfrak{a} \subseteq \mathfrak{a}$. On the other hand, $\mathfrak{a}$ is a right-ideal if for all $r \in R$ we have $\mathfrak{a r} \subseteq \mathfrak{a}$. Furthermore, if $\mathfrak{a}$ is both a left and right ideal, we say that $\mathfrak{a}$ is a two-sided-ideal - or simply an ideal, without qualifiers, which will be the preferred nomenclature.

Remark 8.3.2 (Ideals and 1 ). The only ideal of a ring $R$ containing the unity $1_{R}$ is $R$ itself - thus ideals need not be subrings.

Corollary 8.3.3. If $\phi: R \rightarrow S$ is a ring morphism such that $\operatorname{im} \phi$ is an ideal of $S$, then $\phi$ is surjective.

Proof. Since im $\phi$ is a subring, it contains $1_{S}$ — by Remark 8.3 .2 we have im $\phi=S$. $\quad$.
Corollary 8.3.4. The collection of ideals of a ring is closed under addition and intersection.

Proof. Let $R$ be a ring and $\mathfrak{a}, \mathfrak{b} \subseteq R$ be ideals. Let $a, b \in R$ be any two elements, then $(a+b)(\mathfrak{a}+\mathfrak{b})=(a+b) \mathfrak{a}+(a+b) \mathfrak{b}$ but since $(a+b) \mathfrak{a},(a+b) \mathfrak{b} \subseteq \mathfrak{a}+\mathfrak{b}$, then $(a+b)(\mathfrak{a}+\mathfrak{b}) \subseteq \mathfrak{a}+\mathfrak{b}$ - the same analogous arguments can be used to show that $(\mathfrak{a}+\mathfrak{b})(a+b) \subseteq \mathfrak{a}+\mathfrak{b}$. If $a, b \in \mathfrak{a} \cap \mathfrak{b}$, then $(a b) \mathfrak{a} \cap \mathfrak{b}, \mathfrak{a} \cap \mathfrak{b}(a b) \subseteq \mathfrak{a}, \mathfrak{b}$ thus also contained in $\mathfrak{a} \cap \mathfrak{b}$.

Corollary 8.3.5 (Kernel is an ideal). Given a ring morphism $\phi: R \rightarrow S$, then $\operatorname{ker} \phi$ is a ring ideal of $R$.

Proof. From group theoretic considerations, we already know that ker $\phi$ is a subring of $R$. On the other hand, let $r \in R$ be any element and $a \in \operatorname{ker} \phi$, then $\phi(r a)=\phi(r) \phi(a)=0$ and $\phi(a r)=\phi(a) \phi(r)=0$, thus indeed both $r a$, $a r \in \operatorname{ker} \phi$.

Corollary 8.3.6 (Preimage of ideal is an ideal). Let $\phi: R \rightarrow S$ be a ring morphism and $\mathfrak{a}$ be an ideal of $S$, then the preimage $\phi^{-1}(\mathfrak{a})$ is an ideal of $R$.

Proof. Let $r \in R$ and $a \in \phi^{-1}(\mathfrak{a})$ be any two elements. Notice that $\phi(r a)=\phi(r) \phi(a)$ and $\phi(a r)=\phi(a) \phi(r)$, and since $\phi(a) \in \mathfrak{a}$ then $\phi(r a), \phi(a r) \in \mathfrak{a}$ - therefore $r a, a r \in$ $\phi^{-1}(\mathfrak{a})$.

Remark 8.3.7 (Image of ideals). Although the preimage of ideals is an ideal of the morphism's source, the image of a given ideal need not be an ideal. For instance, let $\mathfrak{a} \subseteq R$ be an ideal of a ring $R$ and $\phi: R \rightarrow S$ be a non-surjective morphism, then there exists $s \in S$ whose preimage is the empty set, therefore $s \phi(\mathfrak{a})$ is not contained in $\phi(\mathfrak{a})$.

Corollary 8.3.8 (Surjective morphisms preserve ideals). If $\phi: R \rightarrow S$ is a surjective ring morphism, then the image of any ideal $\mathfrak{a} \subseteq R$ is an ideal of $S$.

Proof. This is immediate from Remark 8.3.7, for all $s \in S$ we have both $s \phi(\mathfrak{a}), \phi(\mathfrak{a}) s \subseteq$ $\phi(\mathfrak{a})$ because there will exist $r \in R$ whose image is $s$.

Example 8.3.9. For every element $r \in R$ of a ring $R$, the objects $r R$ and $R r$ are, respectively, right and left ideals of $R$. Indeed, given any $a \in R$, we have $(r R) a=r(R a) \subseteq r R$, the analogous being true for the left ideal. If $R$ is commutative, we have $r R=R r$ and such ideal is commonly denoted by $(r)$ - the principal ideal generated by $r$.

Proposition 8.3.10. Let $R$ be a ring and $\left(a_{j}\right)_{j \in J}$ be a collection of ideals of $R$, then the direct sum $\bigoplus_{j \in J} \mathfrak{a}_{j}$ is an ideal of $R$.

Proof. Let $r \in R$ be any element and consider any formal sum $\sum_{j \in J} a_{j} \in \bigoplus_{j \in J} \mathfrak{a}_{j}$ where $a_{j} \neq 0$ for only finitely many $j \in J$, then $r\left(\sum_{j \in J} a_{j}\right)=\sum_{j \in J} r a_{j}$ bus since $r a_{j} \in \mathfrak{a}_{j}$ for each $j \in J$, then $\sum_{j \in J} r a_{j} \in \bigoplus_{j \in J} \mathfrak{a}_{j}$ - the same argument can be used for rightmultiplication.

Lemma 8.3.11. Given a collection $\left(\mathfrak{a}_{j}\right)_{j \in J}$ of ideals of a ring $R$, the ideal $\bigoplus_{j \in J} \mathfrak{a}_{j}$ is the smallest ideal of $R$ containing each ideal $\mathfrak{a}_{j}$ for $j \in J$.

Proof. Let $\mathfrak{b}$ be an ideal of $R$ containing every ideal $\mathfrak{a}_{j}$ for $j \in J$. Then in particular, given any element $\sum_{j \in J} a_{j} \in \bigoplus_{j \in J} \mathfrak{a}_{j}$, since $a_{j} \in \mathfrak{b}$ for every $j \in J$ only finitely many such $a_{j}$ are non-zero, we find that the sum $\sum_{j \in J} a_{j} \in \mathfrak{b}$ - thus indeed $\bigoplus_{j \in J} \mathfrak{a} \subseteq \mathfrak{b}$. $\quad$ a

Definition 8.3.12 (Finitely generated ideals). An ideal a of a commutative ring $R$ is said to be finitely generated if there exists a finite set of elements $A \subseteq R$ such that $\mathfrak{a}=(A)$.

## Division Rings and its Ideals

Proposition 8.3.13 (Units and its ideals). Let $R$ be a ring and $a \in R$. The element $a$ is a left-unit (or right-unit) of $R$ if and only if $R=a R$ (or $R=R a$ )

Proof. We prove the proposition only for left-units, for right-units the proof is completely analogous. Suppose $a$ is a left-unit of $R$ and let $u \in R$ be such that $a u=1$, then $a R$ contains 1 which implies in $a R=R$. Now if $R=a R$, then $1 \in a R$, which means that there must exist $u \in R$ for which $a u=1$, thus $a$ is a left-unit.

Proposition 8.3.14 (Division ring ideals). Let $R$ be a ring. Then $R$ is a division ring if and only if its only left-ideals and right-ideals are $\{0\}$ and $R$.

Proof. Let $R$ be a division ring. Suppose $\mathfrak{a}$ is a left-ideal (or right-ideal) of $R$, then given any $r \in R$ and $a \in \mathfrak{a}$ we have $r a \in \mathfrak{a}$ (or $a r \in \mathfrak{a}$ ) - in particular, if $a \neq 0$, then $a^{-1} \in R$ thus $a^{-1} a=1 \in \mathfrak{a}\left(\right.$ or $a a^{-1}=1 \in \mathfrak{a}$ ), thus $\mathfrak{a}=R$ or $\mathfrak{a}=0$.

Suppose 0 and $R$ are the only left and right ideals. Given any non-zero element $r \in R$, the ideal $r R$ (or $R r$ ) is non-zero, thus must be equal to $R$, which is equivalent of $1 \in r R$ (or $1 \in R r$ ) - therefore $r$ must be a unit of $R$, proving that $R$ is a division ring.

Corollary 8.3.15 (Field ideals). The only left or right ideals of a field $k$ are 0 and $k$.
Proposition 8.3.16. Let $k$ be a field and $R$ be a ring, then any ring morphism $k \rightarrow R$ is injective.

Proof. Let $\phi: k \rightarrow R$ be a ring morphism and let $a, b \in k$ be any two elements such that $\phi(a)=\phi(b)$, then $a-b \in \operatorname{ker} \phi$. Remember that $\operatorname{ker} \phi$ is an ideal of $k$ but this means that $\operatorname{ker} \phi$ is either 0 or $k$ - by Proposition 8.3.14. Since $\phi\left(1_{k}\right)=1_{R}$ for $\phi$ to be a ring morphism, then $\operatorname{ker} \phi \neq k$ and we are left with $\operatorname{ker} \phi=0-$ thus $a=b$.

## Nilpotent Ideal

Definition 8.3.17 (Nilpotent ideal). Let $R$ be a ring. We say that an ideal $\mathfrak{a} \subseteq R$ is a nilpotent ideal if there exists $k \in \mathbf{Z}_{>0}$ such that the product of any $k$ elements of $\mathfrak{a}$ equals zero.

## Nilradicals

Definition 8.3.18 (Nilradical). Given a ring $R$, we define the nilradical of $R$ to be the object $N$ composed of every nilpotent element of $R$.

Corollary 8.3.19. The nilradical $N$ of a commutative ring $R$ is an ideal of $R$.
Proof. Let $a \in R$ and $x \in N$ be any two elements - suppose $n \in \mathbf{Z}_{>0}$ is such that $x^{n}=0$. Since $R$ is commutative then $(a x)^{n}=a^{n} x^{n}=a^{n} \cdot 0=0$, thus $a x \in N$.

Definition 8.3.20 (Reduced ring). A ring $R$ is said to be reduced if it contains no non-zero nilpotent elements

Corollary 8.3.21. Let $R$ be a commutative ring and $N$ be its nilradical. The quotient ring $R / N$ is reduced.

Proof. Suppose $a+N \in R / N$ is a nilpotent element and $a^{n}+N=N$ for some $n \in \mathbf{Z}_{>0}$ - which implies in $a^{n} \in N$ thus there must exist $m \in \mathbf{Z}_{>0}$ such that $\left(a^{n}\right)^{m}=a^{n m}=0$, which implies in $a \in N$ and hence $a+N=N$.
Lemma 8.3.22. A ring $R$ is reduced if and only if for all $r \in R$ such that $r^{2}=0$ implies $r=0$.

Proof. If $R$ is reduced then clearly $r^{2}=0$ implies $r=0$. Let $r \in R$ be any element such that $r^{n}=0$ for some $n \in \mathbf{Z}_{>0}$, then proceed by the following algorithm - start with $r^{n}$, then:

- If $n=1$, return $r$ - since $r=0$.
- If $n$ is odd, $r^{n+1}=0$ and thus $r^{(n+1) / 2} r^{(n+1) / 2}=0$ which by hypothesis implies $r^{(n+1) / 2}=0$. Continue the algorithm for $r^{(n+1) / 2}$.
- If $n$ is even, $r^{n}=r^{n / 2} r^{n / 2}=0$ implies $r^{n / 2}=0$. Continue the algorithm for $r^{n / 2}=0$.

Such algorithm is ensured to terminate and will always result in $r=0$, which implies in $R$ being a reduced ring.

## Quotient Ring

We define now, for every ideal $\mathfrak{a}$ of a given ring $R$, a ring $R / \mathfrak{a}$ together with an additive and multiplicative structure: for every $a+a, b+\mathfrak{a} \in R / a$, we define

$$
\begin{gathered}
(a+\mathfrak{a})+(b+\mathfrak{a}):=(a+b)+\mathfrak{a} \\
(a+\mathfrak{a}) \cdot(b+\mathfrak{a}):=a b+\mathfrak{a} .
\end{gathered}
$$

Let's show that both operations are well defined. Addition is clearly well defined. For the multiplication, suppose $a_{1}+\mathfrak{a}=a_{2}+\mathfrak{a}$ and $b_{1}+\mathfrak{a}=b_{2}+\mathfrak{a}$ - that is, both differences $a_{1}-a_{2}$ and $b_{1}-b_{2} \in \mathfrak{a}$ are elements of $\mathfrak{a}$. Notice that

$$
\begin{aligned}
a_{1} b_{1}-a_{2} b_{2} & =a_{1} b_{1}-a_{2} b_{2}+\left(a_{1} b_{2}-a_{1} b_{2}\right) \\
& =\left(a_{1} b_{1}-a_{1} b_{2}\right)+\left(a_{1} b_{2}-a_{2} b_{2}\right) \\
& =a_{1}\left(b_{1}-b_{2}\right)+\left(a_{1}-a_{2}\right) b_{2}
\end{aligned}
$$

and therefore $a_{1} b_{1}-a_{2} b_{2} \in \mathfrak{a}$, implying in $a_{1} b_{1}+\mathfrak{a}=a_{2} b_{2}+\mathfrak{a}$ - thus multiplication is well defined.

Proposition 8.3.23 (Universal property of quotient rings). Let $R$ be a ring. For every ring $S$ together with a ring morphism $\psi: R \rightarrow S$ and ideal $\mathfrak{a} \subseteq \operatorname{ker} \psi$, there exists a unique ring morphism $\phi: R / \mathfrak{a} \rightarrow S$ such that the following diagram commutes


Proof. Let $S$ be any ring together with a ring morphism $\psi: R \rightarrow S$. Let $\pi: R \rightarrow R /$ a be the canonical projection morphism. Define a map $\phi: R / \mathfrak{a} \rightarrow S$ by sending $r+\mathfrak{a} \mapsto \psi(r)$. We show now that $\phi$ is indeed well defined, consider elements $x, y \in R / \mathfrak{a}$ to be such that $\phi(x)=\phi(y)$, then $\psi(x)=\psi(y)$ which is the same as $\psi(x)-\psi(y)=\psi(x-y)=0$, therefore $x-y \in \operatorname{ker} \psi$, then $x=y$ in $R / a$. Moreover, $\phi$ inherits from $\psi$ the preservation of both additive and multiplicative structures, thus $\phi$ is a morphism of rings and $\phi \pi=\psi$. If $\phi^{\prime}$ is another morphism such that $\phi^{\prime} \pi=\psi$, since $\pi$ is surjective then $\phi \pi=\phi^{\prime} \pi$ implies in $\phi=\phi^{\prime}$, thus $\phi$ is unique.

Corollary 8.3.24 (Every ideal is a kernel). Given a ring $R$, for every ideal $\mathfrak{a} \subseteq R$ is the kernel of some ring morphism $R \rightarrow S$. Therefore an additive subgroup of a ring $R$ is an ideal if and only if it is a kernel of some ring morphism.

Proof. Simply let $S=R / \mathfrak{a}$ and consider the canonical projection $\pi: R \rightarrow R / \mathfrak{a}$, whose kernel is clearly $\mathfrak{a}$.

Definition 8.3.25 (Characteristic). Let $R$ be a ring and $\phi: Z \rightarrow R$ be the unique ring morphism mapping $r \mapsto r \cdot 1_{R}$. We define the characteristic of $R$ to be the non-negative integer $n \in \mathbf{Z}_{>0}$ such that $\operatorname{ker} \phi=n \mathbf{Z}$ - we denote such a property by char $R=n$.

Proposition 8.3.26 (Characteristic of integral domains). The characteristic of an integral domain is either zero or a prime number.

Proof. Let $R$ be an integral domain and let char $R=n$. Suppose that $n$ is non-zero and that there exists an integer $m<n$ dividing $n$ - that is, for some integer $q$ we have $n=q m$. Then in particular $n \cdot 1_{R}=\left(q \cdot 1_{R}\right)\left(m \cdot 1_{R}\right)$ but $n \cdot 1_{R}=0$ and since $q, m<n$ it follows that $q \cdot 1_{R}, m \cdot 1_{R} \neq 0$ in $R$ - thus we obtained a contradiction since $R$ is said to be an integral domain, hence char $R$ is either prime or zero.

Definition 8.3.27 (Boolean ring). A ring $R$ is said to be boolean if for all $r \in R$ we have $r^{2}=r$.

Corollary 8.3.28. A non-zero boolean ring is commutative and has characteristic 2 .
Proof. Let $x \in R$ be any element of a boolean ring $R$. Then we have $x+x=(x+x)^{2}=4 x^{2}$ and $x+x=x^{2}+x^{2}$, therefore, $2 x^{2}=4 x^{2}$ which implies in $x^{2}+x^{2}=0$ but since $x^{2}+x^{2}=x+x$ then we conclude that $x+x=0$ and thus char $R=2$.

If $x, y \in R$ are any two elements, then consider the sum

$$
x+y=(x+y)^{2}=x^{2}+x y+y x+y^{2}=x+x y+y x+y
$$

hence cancelling the common terms we obtain $x y+y x=0$, now since char $R=2$ one can use the fact that $y x+y x=0$ to make

$$
0=x y+y x=x y+y x-(y x+y x)=x y-y x
$$

thus $x y=y x$.

Corollary 8.3.29. If $R$ is a boolean ring and also an integral domain, then there exists a canonical isomorphism of rings $R \simeq \mathbf{Z} / 2 \mathbf{Z}$.

Proof. Let $r \in R$ be any element, since $r^{2}=r$ then $r^{2}-r=r\left(1_{R}-r\right)=0_{R}$ but since $R$ is an integral domain, either $r=0_{R}$ or $1_{R}-r=0_{R}$, that is, $r=1_{R}$. Thus the map $R \rightarrow \mathbf{Z} / 2 \mathbf{Z}$ sending $0_{R} \mapsto[0]_{2}$ and $1_{R} \mapsto[1]_{2}$ is an isomorphism of rings.

## Decompositions

Theorem 8.3.30 (First isomorphism). Every ring morphism $\phi: R \rightarrow S$ can be decomposed into the commutative diagram

where $\bar{\phi}: R / \operatorname{ker} \phi \xrightarrow{\simeq} \operatorname{im} \phi$ is the natural ring morphism induced by $\phi$.
Proof. The morphism $\bar{\phi}$ is obtained by the quotient ring universal property - that is, $\bar{\phi}(r)=\phi(r)+\operatorname{ker} \phi$, which is a morphism $R / \operatorname{ker} \phi \rightarrow S$. Restricting the codomain of $\bar{\phi}$ we obtain the claimed isomorphism. The rest follows trivially.

Corollary 8.3.31. Let $\phi: R \rightarrow S$ be a surjective ring morphism, then there exists a natural isomorphism

$$
S \simeq R / \operatorname{ker} \phi
$$

Proof. Indeed, if $\phi$ is surjective, then $\operatorname{im} \phi=S$ and by the first isomorphism theorem we obtain the natural isomorphism $S \simeq R / \operatorname{ker} \phi$.

Proposition 8.3.32 (Ideal of a quotient). Let $R$ be a ring and $\mathfrak{a} \subseteq R$ be an ideal. If $\mathfrak{b}$ is an ideal of $R$ and $\mathfrak{a} \subseteq \mathfrak{b}$, then $\mathfrak{b} / \mathfrak{a}$ is an ideal of $R / \mathfrak{a}$ and there exists a natural isomorphism

$$
\frac{R / \mathfrak{a}}{\mathfrak{b} / \mathfrak{a}} \simeq R / \mathfrak{b}
$$

Proof. By Proposition 8.3.23, let $\phi: R / \mathfrak{a} \rightarrow R / \mathfrak{b}$ be the morphism making the following diagram commute


That is, $\phi$ is defined by mapping $r+\mathfrak{a} \mapsto r+\mathfrak{b}$. Notice that the kernel of $\phi$ is composed of those elements $r+\mathfrak{a} \in R / \mathfrak{a}$ for which $\phi(r+\mathfrak{a})=\mathfrak{b}$, that is, $r \in \mathfrak{b}$ for $r+\mathfrak{b}=\mathfrak{b}-$ thus $\operatorname{ker} \phi=\mathfrak{b} / \mathfrak{a}$, which is ensured to be an ideal. Since $\phi$ is surjective, by means of Corollary 8.3.31 we obtain a natural isomorphism $R / \mathfrak{b} \simeq \frac{R / \mathfrak{a}}{\mathfrak{b} / \mathfrak{a}}$.

Corollary 8.3.33. Any surjective morphism $\phi: R \rightarrow S$ can be bijectively identified as a canonical projection $R \rightarrow R / \operatorname{ker} \phi$.

Example 8.3.34. If $R$ is a commutative ring and we consider principal ideals ( $a$ ) and (b) of $R$. If $[b] \in R /(a)$ is the class of $b$, then we have $([b])=(a, b) /(a)$. By Proposition 8.3.32 we find a canonical isomorphism

$$
\frac{R /(a)}{([b])} \simeq R /(a, b)
$$

## Generation of Ideals

Definition 8.3.35 (Noetherian ring). A commutative ring $R$ is said to be Noetherian if every ideal of $R$ is finitely generated.

Proposition 8.3.36 (Image of Noetherian ring). Let $R$ be a Noetherian ring and $S$ be a ring. If there exists a surjective ring morphism $R \rightarrow S$ then $S$ is Noetherian.

Proof. Suppose there exists $\phi: R \rightarrow S$, a surjective ring morphism. Let $\mathfrak{s}$ be any ideal of $S$, since the preimage of an ideal is an ideal, then $\phi^{-1}(\mathfrak{s})$ is an ideal of $R$ - thus finitely generated. Now, since $\phi$ is surjective, $\phi^{-1}(\mathfrak{s})$ is non-empty and there exists a finite set $A \subseteq \phi^{-1}(\mathfrak{s})$ such that $\phi^{-1}(\mathfrak{s})=(A)$, since $R$ is Noetherian. Therefore $\phi(A)=B \subseteq \mathfrak{s}$ is a finite set, hence $\phi((A))=(B)=\mathfrak{s}$ - proving that $\mathfrak{s}$ is finitely generated.

Definition 8.3.37 (Principal ideal domain). An integral domain $R$ is said to be a principal ideal domain (which we'll shortly name PID) if every ideal of $R$ is principal.

Example 8.3.38. The ring of integers $\mathbf{Z}$ is a PID. Indeed, if $\mathfrak{a}$ is an ideal of $\mathbf{Z}$, then there exists $n \in \mathbf{Z}$ for which $\mathfrak{a}$ is a subgroup of $n \mathbf{Z}=(n)$, thus $\mathfrak{a}$ itself is principal.

Notice also that, given any two integers $m, n \in \mathbf{Z}$, if $d:=\operatorname{gcd}(m, n)$ then $m, n \in(d)$, which implies that $(m, n) \subseteq(d)$. Moreover, from Bézout's identity ${ }^{1}$, we have the existence of $a, b \in \mathbf{Z}$ for which $a m+b n=d$, thus $d \in(m, n)$ - proving that $(m, n)=(d)$.

Example 8.3.39. The ring $\mathbf{Z}[x]$ is not a PID.
We consider the ideal $(2, x)$ and show that it isn't principal. Suppose there exists $f(x) \in \mathbf{Z}[x]$ for which $(f(x))=(2, x)$, so that there exists $q(x) \in \mathbf{Z}[x]$ such that $q(x) f(x)=2$. Notice however that since $\mathbf{Z}$ is an integral domain, the product of the leading term coefficients of $f(x)$ and $q(x)$ is necessarily non-zero (for non-zero polynomials), thus $\operatorname{deg} f(x) g(x)=\operatorname{deg} f(x)+\operatorname{deg} g(x)$. Notice that $\operatorname{deg} q(x) f(x)=0$

[^11]thus $\operatorname{deg} f(x)=0$ and this can't be the case since we would have $x \notin(f(x))$ - this is a contradiction, such polynomial $f(x)$ cannot exist and therefore $(2, x)$ isn't principal.

Proposition 8.3.40. Given a field $k$, the ring of polynomials $k[x]$ is a PID.
Proof. The zero ideal is obviously principal, thus let $\mathfrak{a}$ be any non-zero proper ideal of $k[x]$ and, by the well-ordering of the positive degree polynomials in $k[x]$, let $f(x) \in \mathfrak{a}$ be a non-zero monic polynomial with least degree. We can indeed be certain that we can choose an $f(x)$, because every non-zero coefficient of a polynomial of $k[x]$ is a unity - since $\mathfrak{a}$ is non-zero, we are ensured that there will exist a polynomial with at least one non-zero coefficient.

Let $g(x) \in k[x]$ be any polynomial and, since $k$ is a field, let $q(x), r(x) \in k[x]$ be polynomials such that $g(x)=q(x) f(x)+r(x)$, where $\operatorname{deg} r<\operatorname{deg} f(x)$. Notice then that the last condition implies in $\operatorname{deg} r(x) \leqslant 0$, otherwise $f(x)$ wouldn't be the monic polynomial with least degree among polynomials of positive degree in $k[x]$. Now, if $r(x)=0$, then $g(x)=q(x) f(x)$ and $g(x) \in(f(x))$ - on the other hand, if $r(x)=a$ for some $a \in k$, then $a \in \mathfrak{a}$ but since $k$ is a field, this implies in $1 \in \mathfrak{a}$ thus $\mathfrak{a}=k[x]$ and hence a contradiction to the hypothesis that $\mathfrak{a}$ is proper. Assuming the latter does not hold, we find that $(f(x))=\mathfrak{a}$ and hence $k[x]$ is a PID.

Definition 8.3.41 (Product of ideals). Given a ring $R$ and ideals $\mathfrak{a}$ and $\mathfrak{b}$ of $R$, we denote by $\mathfrak{a b}$ the ideal generated by all products $a b$ for $a \in \mathfrak{a}$ and $b \in \mathfrak{b}$.

Lemma 8.3.42. Let $\mathfrak{a}$ and $\mathfrak{b}$ be ideals of a commutative ring $R$. If either one of the following properties is true:
(a) The ideals $\mathfrak{a}$ and $\mathfrak{b}$ are comaximal, that is, $\mathfrak{a}+\mathfrak{b}=R$.
(b) The quotient $R /(\mathfrak{a b})$ is a reduced ring.

Then it follows that

$$
\mathfrak{a b}=\mathfrak{a} \cap \mathfrak{b}
$$

Proof. If $a b \in \mathfrak{a b}$ is any element, then $a, b \in \mathfrak{a} \cap \mathfrak{b}$ and hence $a b \in \mathfrak{a} \cap \mathfrak{b}$ - thus in either cases we have $\mathfrak{a b} \subseteq \mathfrak{a} \cap \mathfrak{b}$. We now prove the other side of the inclusion for each of the properties - let $\ell \in \mathfrak{a} \cap \mathfrak{b}$ be any element:
(a) Since $\mathfrak{a}+\mathfrak{b}=R$, then there are $a \in \mathfrak{a}$ and $b \in \mathfrak{b}$ such that $a+b=1$. Moreover, since $\ell \in \mathfrak{a} \cap \mathfrak{b}$, it follows that $a \ell+b \ell=\ell$ is an element of $\mathfrak{a b}$, thus $\mathfrak{a} \cap \mathfrak{b} \subseteq \mathfrak{a b}$.
(b) Notice that we have $\ell^{2} \in \mathfrak{a b}$, thus $\ell^{2}+\mathfrak{a b}=\mathfrak{a b}$. Since $R /(\mathfrak{a b})$ is reduced, it follows that $\ell+\mathfrak{a b}=\mathfrak{a b}$ (see Lemma 8.3.22) - thus $\ell \in \mathfrak{a b}$ and $\mathfrak{a} \cap \mathfrak{b} \subseteq \mathfrak{a b}$.

Lemma 8.3.43. Let $R$ be a ring and $f(x) \in R[x]$ be a monomial. Then $f(x)$ is not a left or right zero-divisor.

Proof. Since $f(x)$ is is a monomial and there is no element $r \in R$ other than zero for which $r \cdot 1_{R}=0$ (where $1_{R}$ is the coefficient of the leading term of $f(x)$ ), then it follows that the only polynomial $z(x) \in R[x]$ for which $f(x) z(x)=0$ (or $z(x) f(x)=0)$ is the zero polynomial $z(x)=0$.

Lemma 8.3.44 (Degree of the product of polynomials). If $f(x) \in R[x]$ is monic a polynomial and $R$ is a ring, then for every $g(x) \in R[x]$ we have

$$
\operatorname{deg}(f(x) g(x))=\operatorname{deg} f(x)+\operatorname{deg} g(x)
$$

One can extend the euclidean division algorithm to the ring of polynomials, where we'll be able to divide polynomials by monomials.

Lemma 8.3.45 (Division of polynomials). Given a monic polynomial $f(x) \in R[x]$, for some ring $R$, then for all $g(x) \in R[x]$ there exists two polynomials $q(x), r(x) \in R[x]$ for which

$$
g(x)=f(x) q(x)+r(x)
$$

and $\operatorname{deg} r(x)<\operatorname{deg} f(x)$. Moreover, such polynomials $q(x)$ and $r(x)$ are unique.
Proof. Indeed, if $q^{\prime}(x), r^{\prime}(x) \in R[x]$ are polynomials - and, for the sake of contradiction, distinct from the respective $q(x)$ and $r(x)$ - for which $g(x)=f(x) q^{\prime}(x)+r^{\prime}(x)$, and $\operatorname{deg} r^{\prime}(x)<\operatorname{deg} f(x)$, we find that $f(x) q(x)+r(x)=f(x) q^{\prime}(x)+r^{\prime}(x)$ and thus

$$
\begin{equation*}
r(x)-r^{\prime}(x)=f(x)\left(q^{\prime}(x)-q(x)\right) \tag{8.1}
\end{equation*}
$$

Notice however that by hypothesis $\operatorname{deg} r(x), \operatorname{deg} r^{\prime}(x)<\operatorname{deg} f(x)$, thus $\operatorname{deg}(r(x)-$ $\left.r^{\prime}(x)\right) \leqslant \max \left(\operatorname{deg} r(x), \operatorname{deg} r^{\prime}(x)\right)<\operatorname{deg} f(x)$ — which is in contradiction with Eq. (8.1) since for such equation to be true we should have $\operatorname{deg}\left(r(x)-r^{\prime}(x)\right)=\operatorname{deg}\left(f(x)\left(q^{\prime}(x)-\right.\right.$ $q(x)))=\operatorname{deg} f(x)+\operatorname{deg}\left(q^{\prime}(x)-q(x)\right)$. Therefore $r(x)-r^{\prime}(x)$ is necessarily the zero polynomial and $f(x) q(x)=f(x) q^{\prime}(x)$, now since $f(x)$ is a monomial and by Lemma 8.3.43 is not a zero-divisor, we use Proposition 8.1.14 to conclude that $q(x)=q^{\prime}(x)$.

In what follows, one should regard $R^{\oplus d}$, for a ring $R$, to be the ring of polynomials with degree less than or equal to $d$ - this observation is based on the fact that one can map injectively

$$
\begin{equation*}
\Psi: R^{\oplus d} \mapsto R[x] \text { sending }\left(r_{0}, \ldots, r_{d-1}\right) \mapsto \sum_{j=0}^{d-1} r_{j} x^{j} \tag{8.2}
\end{equation*}
$$

which is a morphism of abelian groups. Therefore, the restriction of the codomain to the image of such map induces an isomorphism of abelian groups

$$
\begin{equation*}
R^{\oplus} \simeq \operatorname{im} \Psi \subseteq R[x] . \tag{8.3}
\end{equation*}
$$

Proposition 8.3.46. Let $R$ be a commutative ring, and $f(x) \in R[x]$ a monic polynomial with degree $d \in \mathbf{Z}_{>0}$. Define a map $\phi: R[x] \rightarrow R^{\oplus d}$ sending $g(x) \mapsto\left(r_{0}, \ldots, r_{d-1}\right)$, where $r(x):=\Psi\left(r_{0}, \ldots, r_{d-1}\right) \in R[x]$ is the remainder of the division of $g(x)$ by $f(x)$. Such map $\phi$ induces a natural isomorphism of abelian groups

$$
R[x] /(f(x)) \simeq R^{\oplus d}
$$

Proof. Notice that every polynomial $g(x)$ can be divided by a monic polynomial $f(x)$ and the remainder $r(x)$ of such division is unique by Lemma 8.3.45. Moreover, the isomorphism in Eq. (8.3) is a right inverse of $\phi$, thus $\phi$ is surjective. Moreover, $\operatorname{ker} \phi=$ $(f(x))$ since every polynomial in the principal ideal $(f(x))$ is divisible by $f(x)$ and hence has a zero remainder when divided by $f(x)$.

Let's check that $\phi$ is indeed a morphism of abelian groups. Let $g(x), g^{\prime}(x) \in R[x]$ be two polynomials and let $q(x), r(x) \in R[x]$ and $q^{\prime}(x), r^{\prime}(x) \in R[x]$ be their respective pair of quotient and remainder in the division by $f(x)$ - where $\operatorname{deg} r(x), \operatorname{deg} r^{\prime}(x)<$ $\operatorname{deg} f(x)$. Therefore, the sum of $g(x)$ with $g^{\prime}(x)$ can be written as

$$
g(x)+g^{\prime}(x)=f(x)\left(q(x)+q^{\prime}(x)\right)+\left(r(x)+r^{\prime}(x)\right) .
$$

Since $\operatorname{deg}\left(r(x)+r^{\prime}(x)\right) \leqslant \max \left(\operatorname{deg} r(x), \operatorname{deg} r^{\prime}(x)\right)<\operatorname{deg} f(x)$, then $r(x)+r^{\prime}(x)$ is the remainder of the division of $g(x)+g^{\prime}(x)$ by $f(x)$. With this in our hands we can rightly see that

$$
\begin{aligned}
\phi\left(g(x)+g^{\prime}(x)\right) & =\left(r_{0}+r_{0}^{\prime}, \ldots, r_{d-1}+r_{d-1}^{\prime}\right) \\
& =\left(r_{0}, \ldots, r_{d-1}\right)+\left(r_{0}^{\prime}, \ldots, r_{d-1}^{\prime}\right) \\
& =\phi(g(x))+\phi\left(g^{\prime}(x)\right),
\end{aligned}
$$

where $\Psi\left(r_{0}, \ldots, r_{d-1}\right):=r(x)$ and $\Psi\left(r_{0}^{\prime}, \ldots, r_{d-1}^{\prime}\right):=r^{\prime}(x)$ - thus $\phi$ is a morphism of abelian groups. By Proposition 7.4.16 we find

$$
R[x] / \operatorname{ker} \phi=R[x] /(f(x)) \simeq \operatorname{im} \phi=R^{\oplus d} .
$$

Example 8.3.47 (Evaluation). Let $R$ be a commutative ring. The evaluation morphism eval $_{a}: R[x] \rightarrow R$ (see Proposition 8.2.9), mapping $f(x) \mapsto f(a)$, induces a natural isomorphism of abelian groups

$$
R[x] /(x-a) \simeq R
$$

Such isomorphism, name it $\phi$, is given by the map $g(x)+(x-a) \mapsto r$, where $r \in R$ is the remainder of the division of $g(x)$ by $x-a$.

First of all, if $f(x) \in R[x]$ then the division by $x-a$ yields $q(x) \in R[x]$ and $r \in R$ such that $f(x)=(x-a) q(x)+r$ since the degree of the remainder should be less than $\operatorname{deg}(x-a)=1$. Therefore, evaluating $\operatorname{eval}_{a}(f(x))=(a-a) q(x)+r=r$ implies in $f(a)=r$ - thus $f(x) \in \operatorname{ker} \phi$ if and only if $f(x) \in(x-a)$, hence ker $\phi=(x-a)$. Moreover, we can now rethink of the $\phi$ as the mapping $f(x)+(x-a) \mapsto f(a)$.

If $u(x)=u(a)+(x-a)$ and $v(x)=v(a)+(x-a)$ are any two polynomials in $R[x] /(x-a)$, one finds that $u(x)+v(x)=(u(a)+v(a))+(x-a)$ and therefore $\phi(u(x)+v(x))=$ $u(a)+v(a)=\phi(u(x))+\phi(v(x))$, thus $\phi$ is a morphism of abelian groups. Therefore, by means of Corollary 8.3.31 we find that $\phi$ is an isomorphism.

Example 8.3.48 (Constructing C). Consider, as in Proposition 8.3.46, the map $\phi: \mathbf{R}[x] \rightarrow$ $\mathbf{R} \oplus \mathbf{R}$ sending $f(x) \mapsto\left(r_{0}, r_{1}\right)$, where $r(x)=r_{0}+r_{1} x$ is the remainder of the division of $f(x)$ by the polynomial $x^{2}+1 \in \mathbf{R}[x]$.

For any element $f(x):=a_{0}+a_{1} x \in \mathbf{R}[x]$ we have $\operatorname{deg} f(x)<\operatorname{deg}\left(x^{2}+1\right)=2$, then $\phi(f(x))=\left(a_{0}, a_{1}\right)$. Hence, given another $g(x)=b_{0}+b_{1} x \in \mathbf{R}[x]$, one has that

$$
\begin{aligned}
f(x) g(x) & =\left(a_{0}+a_{1} x\right)\left(b_{0}+b_{1} x\right) \\
& =a_{0} b_{0}+\left(a_{0} b_{1}+a_{1} b_{0}\right) x+a_{1} b_{1} x^{2} \\
& =\left(x^{2}+1\right) a_{1} b_{1}+\left(\left(a_{0} b_{0}-a_{1} b_{1}\right)+\left(a_{0} b_{1}+a_{1} b_{0}\right) x\right) .
\end{aligned}
$$

Thus $\phi(f(x) g(x))=\left(a_{0} b_{0}-a_{1} b_{1}, a_{0} b_{1}+a_{1} b_{0}\right)$, this induces a multiplicative structure $\because(\mathbf{R} \oplus \mathbf{R})^{2} \rightarrow \mathbf{R} \oplus \mathbf{R}$ defined by

$$
\left(a_{0}, a_{1}\right) \cdot\left(b_{0}, b_{1}\right):=\left(a_{0} b_{0}-a_{1} b_{1}, a_{0} b_{1}+a_{1} b_{0}\right)
$$

Now, since $(\mathbf{R} \oplus \mathbf{R}, \cdot) \simeq \mathbf{C}$, by identifying $(a, b) \mapsto a+b i$, one obtains the following natural isomorphism

$$
\mathbf{R}[x] /\left(x^{2}+1\right) \simeq \mathbf{C}
$$

hence we've constructed the complex numbers out of the ring $\mathbf{R}[x]$.
Example 8.3.49. Let $R$ be a commutative ring and elements $a_{1}, \ldots, a_{n} \in R$. Then there exists a canonical isomorphism of abelian groups

$$
\frac{R\left[x_{1}, \ldots, x_{n}\right]}{\left(x_{1}-a_{1}, \ldots, x_{n}-a_{n}\right)} \simeq R
$$

Denote $S:=\frac{R\left[x_{1}, \ldots, x_{n}\right]}{\left(x_{1}-a_{1}, \ldots, x_{n}-a_{n}\right)}$. Consider the evaluation map eval: $S \rightarrow R$ given by $\left[p\left(x_{1}, \ldots, x_{n}\right)\right] \mapsto p\left(a_{1}, \ldots, a_{n}\right)$, which is clearly a group morphism. Moreover, the map $\rho: R \rightarrow S$ mapping $r \mapsto[r]$ is both a group morphism and right inverse of the evaluation map, eval $\circ \rho=\operatorname{id}_{R}$. Since for all $1 \leqslant j \leqslant n$ we have $\left[x_{j}\right]=\left[a_{j}\right]$ as classes in the quotient ring $S$, we find that, for any given $p\left(x_{1}, \ldots, x_{n}\right) \in R[x]$

$$
\left[p\left(x_{1}, \ldots, x_{n}\right)\right]=\left[p\left(a_{1}, \ldots, a_{n}\right)\right],
$$

implying in $\rho \circ$ eval $=\mathrm{id}_{S}$.

## Prime \& Maximal Ideals

Definition 8.3.50 (Prime and maximal ideal). Let $R$ be a commutative ring and $\mathfrak{a}$ be a proper ideal of $R$. We define the following:
(a) If $R / \mathfrak{a}$ is an integral domain, we call $\mathfrak{a}$ a prime ideal.
(b) If $R / \mathfrak{a}$ is a field, we call $\mathfrak{a}$ a maximal ideal.

Example 8.3.51. Given a commutative ring $R$ and any monomial $x-a \in R[x]$, then:

- The ideal $(x-a)$ is prime if and only if $R$ is an integral domain.
- The ideal $(x-a)$ is maximal if and only if $R$ is a field.

Such propositions are direct implications of the isomorphism $R[x] /(x-a) \simeq R$ given by Proposition 8.3.46.

Some may want to avoid the machinery of quotient rings when defining prime and maximal ideals. Such ambition is palpable, the following proposition exposes this old fashioned alternative.

Proposition 8.3.52. Let $\mathfrak{a}$ be a proper ideal of a commutative ring $R$ - then:
(a) The ideal $\mathfrak{a}$ is prime if and only if for every given $a, b \in R$ such that $a b \in \mathfrak{a}$, we have either $a \in \mathfrak{a}$ or $b \in \mathfrak{a}$.
(b) The ideal $\mathfrak{a}$ is maximal if and only the only ideals that contain $\mathfrak{a}$ is either $\mathfrak{a}$ itself or $R$.

Proof. 1. The quotient ring $R / \mathfrak{a}$ is an integral domain if and only if for all $a+\mathfrak{a}, b+\mathfrak{a} \in$ $R / \mathfrak{a}$ such that $(a+\mathfrak{a})(b+\mathfrak{a})=\mathfrak{a}$ we have $a+\mathfrak{a}=\mathfrak{a}$ or $b+\mathfrak{a}=\mathfrak{a}$ - which is equivalent to the requirement that either $a \in \mathfrak{a}$ or $b \in \mathfrak{b}$.
2. The quotient ring $R / \mathfrak{a}$ is a field if and only if $\mathfrak{a}$ and $R / \mathfrak{a}$ are the only ideals of $R / \mathfrak{a}$ - that is, if $\mathfrak{b}$ is an ideal of $R$ with $\mathfrak{a} \subseteq \mathfrak{b}$, then $\mathfrak{b}$ is also an ideal of $R / \mathfrak{a}$ (see Proposition 8.3.32), thus either $\mathfrak{b}=\mathfrak{a}$ or $\mathfrak{b}=R / \mathfrak{a}$.

Corollary 8.3.53. Maximal ideals are prime ideals.
Proof. Let $R$ be a commutative ring and $\mathfrak{m}$ be maximal. Let $a, b \in R$ be any elements such that $a b \in \mathfrak{m}$, then consider the principal ideals $(a)$ and $(b)$. Since $\mathfrak{m}$ is maximal, such principal ideals can be either $R$ or $\mathfrak{m}$ - for the former case, one of them equals $1_{R}$, thus the other is contained in $\mathfrak{m}$, on the other hand, for the latter case the element whose ideal equal to $\mathfrak{m}$ belongs to $\mathfrak{m}$. This shows that either $a \in \mathfrak{m}$ or $b \in \mathfrak{m}$.

Proposition 8.3.54. Let $\mathfrak{a}$ be an ideal of a commutative ring $R$. If $R / a$ is a finite ring, then $\mathfrak{a}$ is prime if and only if it is maximal.

Proof. Since a finite commutative ring is an integral domain if and only if it is a field, the proposition follows - see Proposition 8.1.31.

Example 8.3.55. A good example of Proposition 8.3.54 in action is the case of the ring $\mathbf{Z}$ and the principal ideals $(n)$ for $n \in \mathbf{Z}_{>0}$. Since $\mathbf{Z} /(n)$ is always finite, $(n)$ is a prime ideal if and only if $(n)$ is maximal - on the other hand, $(n)$ is maximal if and only if in the case where $n$ is a prime number.

Definition 8.3.56 (Ring spectrum). Given a commutative ring $R$, we define the following:
(a) The collection of all prime ideals of $R$ is called the prime spectrum of $R$ and is denoted Spec $R$.
(b) The collection of all maximal ideals of $R$ is called the maximal spectrum of $R$ and is denoted Spec $_{\mathrm{m}} R$.

Proposition 8.3.57 (Prime \& maximal ideals in PIDs). Let $R$ be a PID, and let $\mathfrak{a}$ be a non-zero ideal of $R$. Then the ideal $\mathfrak{a}$ is prime if and only if it is maximal.

Proof. Suppose $\mathfrak{a}$ is prime. Since $R$ is PID, we can be certain that there exists a non-zero element $a \in R$ such that $\mathfrak{a}=(a)$. Now, consider $\mathfrak{b}:=(b)$ to be any ideal of $R$ containing $\mathfrak{b}$, then in particular there exists $q \in \mathbf{R}$ such that $a=b q$ - since $\mathfrak{a}$ is prime and $b q \in \mathfrak{a}$ then either $b \in \mathfrak{a}$ or $q \in \mathfrak{a}$ :

- For the case where $b \in \mathfrak{a}$ then $\mathfrak{b} \subseteq \mathfrak{a}$ - thus from hypothesis that $\mathfrak{b}$ contained $\mathfrak{a}$ we conclude that $\mathfrak{b}=\mathfrak{a}$.
- If $q \in \mathfrak{a}$, then there exists $d \in R$ such that $q=a d$, therefore $a=b q=b(a d)=$ $b(d a)=(b d) a$ - since a PID is an integral domain and $a \neq 0$, one can cancel $a$ from both sides to obtain $b d=1$, thus $b$ is a unit element of $R$. In particular $b d=1 \in \mathfrak{b}$, thus $\mathfrak{b}=R$.

By Proposition 8.3.52 we conclude that $\mathfrak{a}$ is maximal.
Example 8.3.58 (Prime ideals of $k[x]$ are maximal). Let $k$ be a field. Then non-zero prime ideals of $k[x]$ are maximal. Indeed, from Proposition 8.3.40 we know that $k[x]$ is PID, thus this is all but a consequence of Proposition 8.3.57.

Proposition 8.3.59 (Maximal ideals in algebraically closed fields). Let $k$ be an algebraically closed field and $\mathfrak{a}$ be an ideal of $k[x]$. Then $\mathfrak{a}$ is maximal if and only if there exists some $a \in k$ for which $\mathfrak{a}=(x-a)$.

Proof. From Example 8.3 .51 we know that $(x-a)$ is a maximal ideal of $k[x]$, since $k$ is a field.

Suppose $\mathfrak{a}$ is maximal, since $k[x]$ is a PID, it follows that there exists an $f(x) \in k[x]$ for which $\mathfrak{a}=(f(x))$. Moreover, since $k$ is algebraically closed, there exists $a \in R$ for which $f(a)=0$, therefore $f(x) \in(x-a)$ and then $\mathfrak{a} \subseteq(x-a)$. From the hypothesis that $\mathfrak{a}$ is maximal, we conclude that either $(x-a)=\mathfrak{a}$ or $(x-a)=k[x]$ - notice that the latter cannot be the case since the collection of constant polynomials is not contained in $(x-a)$, thus $\mathfrak{a}=(x-a)$.

Example 8.3.60. Let $K$ be a compact topological space and $C(K)$ be the ring of continuous maps $K \rightarrow \mathbf{R}$, with addition and multiplication defined point-wise. We have the following facts:
(a) For every $p \in K$, the collection $\mathfrak{m}_{p}:=\{f \in C(K): f(p)=0\}$ is a maximal ideal of $R$.
(b) If $f_{1}, \ldots, f_{n} \in C(K)$ have no common zeros, then $\left(f_{1}, \ldots, f_{n}\right)=R$.
(c) For every maximal ideal $\mathfrak{m} \in \operatorname{Spec}_{\mathrm{m}} C(K)$, there exists a point $p \in K$ for which $\mathfrak{m}=\mathfrak{m}_{p}$.
(d) If in addition $K$ is Hausdorff, then there exists a bijective set-function $K \xrightarrow{\simeq}$ Spec $_{m} C(K)$ given by $p \mapsto \mathfrak{m}_{p}$.

Proof. We now prove each one of these statements.
(a) Consider any map $f+\mathfrak{m}_{p} \in C(K) / \mathfrak{m}_{p}$ such that $f(p) \neq a$ for some $a \neq 0$ and therefore $f \notin \mathfrak{m}_{p}$. Notice that the unit of $C(K) / \mathfrak{m}_{p}$ is composed of every element $g \in C(K)$ such that $g(p)=1$. In particular, since $a$ is non-zero, then there exists an inverse $a^{-1} \in \mathbf{R}$. If $g \in C(K)$ is any map assuming $g(p)=a^{-1}$, then $\left(f+\mathfrak{m}_{p}\right)\left(g+\mathfrak{m}_{p}\right)=1+\mathfrak{m}_{p}$.
(b) Since $f_{1}, \ldots, f_{n}$ share no zeros, the function $f:=\sum_{j=1}^{n} f_{j}^{2}$ must be strictly positive, therefore one can define a continuous map $g \in C(K)$ as $g(x):=\frac{1}{f(x)}$ for all $x \in K$. Therefore $g f=1$, which implies in $1 \in\left(f_{1}, \ldots, f_{n}\right)$ - where we denote by 1 the constant map assuming value 1.
(c) Let $\mathfrak{m}$ be maximal and suppose, for the sake of contradiction, that for all $p \in K$ we have $\mathfrak{m} \neq \mathfrak{m}_{p}$. Fix any $p \in K$ and let $f_{p} \in \mathfrak{m}$ be such that $f_{p}(p) \neq 0$. By the continuity of $f_{p}$ there exists a neighbourhood $U_{p} \subseteq K$ of $p$ such that $f(x) \neq 0$ for all $x \in U_{p}$. Let $\mathcal{U}:=\left\{U_{p}\right\}_{p \in K}$ be an open cover of $K$, where each neighbourhood $U_{p}$ is associated with a map $f_{p}$ as described above. Since $K$ is compact, there exists a finite collection $U_{p_{1}}, \ldots, U_{p_{n}} \in \mathcal{U}$ covering $K$. From construction, we obtain $\left(f_{p_{1}}, \ldots, f_{p_{n}}\right) \subseteq \mathfrak{m}$. Notice that the maps $f_{p_{j}}$ share no common zero - if $q \in K$ is such that $f_{p_{j}}(q)=0$ for each $1 \leqslant j \leqslant n$, then $q \notin \bigcup_{j=1}^{n} U_{p_{j}}$, which is a contradiction - therefore $1 \in \mathfrak{m}$ and $\mathfrak{m}=C(K)$. We conclude that there must exist $p \in K$ such that $\mathfrak{m}=\mathfrak{m}_{p}$.

## Prove last item, requires Urysohn's lemma

Definition 8.3.61 (Krull dimension). The Krull dimension of a commutative ring $R$ is defined as the length of the longest chain of prime ideals in $R$.

### 8.4 Modules Over Rings

## Modules

Definition 8.4.1 (Ring action). Let $R$ be a ring and $A$ an abelian group. A left- $R$-action on $A$ is given a ring morphism

$$
\mu: R \rightarrow \operatorname{End}_{\mathrm{Ab}}(M)
$$

Less compactly, given $r, s \in R$ and a group endomorphism $\phi: A \rightarrow A$, the ring morphism satisfies, for every $a, b \in A$ :
(a) $\mu(r)(a+b)=\mu(r)(a)+\mu(r)(b)$.
(b) $\mu(r+s)(a)=\mu(r)(a)+\mu(s)(a)$.
(c) $\mu(r s)(a)=\mu(r)(\mu(s)(a))$.
(d) $\mu(1)(a)=a$.

Definition 8.4.2 (Module). Given a ring $R$, a left- $R$-module is an abelian group $M$ endowed with a left-R-action $R \times M \rightarrow M$ mapping $(r, m) \mapsto r m$ satisfying the following properties, for every $r, s \in R$ and $m, n \in M$ :
(a) $r(m+n)=r m+r n$.
(b) $(r+s) m=r m+s m$.
(c) $(r s) m=r(s m)$.
(d) $1 m=m$.

Right- $R$-modules are defined completely analogous, with a right- $R$-action.
Definition 8.4.3 (Opposite ring). Given a ring $(R,+, \cdot)$, we define its opposite ring $R^{\mathrm{op}}$ to be ring inheriting the elements and additive structure of $R$, while its multiplicative structure is given by a map $*: R \times R \rightarrow R$ defined by $a * b:=b a \in R$.

Corollary 8.4.4. The identity map id: $R \rightarrow R^{\mathrm{op}}$ is an isomorphism of rings if and only if $R$ is commutative.

Proof. If $R$ is commutative, clearly the identity map is an isomorphism. Conversely, if id is a morphism of rings, then $\operatorname{id}(r s)=\operatorname{id}(r) * \operatorname{id}(s)=\operatorname{id}(s) \operatorname{id}(r)=s r$ but $\operatorname{id}(r s)=r s$ thus $r s=s r$.

Example 8.4.5. The ring of real $n$-square matrices, $M_{n}(\mathbf{R})$, is isomorphic as a ring to its opposite ring $M_{n}(\mathbf{R})^{\mathrm{op}}$.

Indeed, if $\phi: M_{n}(\mathbf{R}) \rightarrow M_{n}(R)^{\mathrm{op}}$ is the map sending $A \mapsto A^{T}$ then, given $A, B \in$ $M_{n}(\mathbf{R})$, we have

$$
\phi(A B)=(A B)^{T}=B^{T} A^{T}=A^{T} * B^{T}=\phi(A) * \phi(B) .
$$

That is, $\phi$ is a morphism of rings. Injectivity and surjectivity are clear, thus $\phi$ is an isomorphism of rings.

Example 8.4.6. For a commutative ring $R$, left- $R$-modules can be bijectively assigned to a right- $R$-modules. To see that given an abelian group $M$, let ${ }_{R} M$ be a left- $R$-module structure on $M$, we proceed by constructing a bijection ${ }_{R} M \stackrel{\simeq}{\leftrightharpoons} M_{R}$, where $M_{R}$ denotes a right- $R$-module on $M$. For every element $r \in R$ and $m \in M$, map $r m \mapsto m r$. Since $R$ is commutative, indeed

$$
(r s) m=r(s m) \longmapsto m(r s)=m(s r)=(m s) r
$$

translating the multiplication by $r s$ from a left-module to a right-module accordingly.

## Vector Spaces

Example 8.4.7 (Vector spaces). A module over a field $k$ is nothing more than a $k$-vector space, thus $k$-Mod $=\operatorname{Vect}_{k}$.

Example 8.4.8 $(k[x]$-module structure on $V)$. Let $k$ be a field, $V$ be a $k$-vector space, and $\phi: V \rightarrow V$ be an endomorphism. We consider the ring action $\mu: k \hookrightarrow$ End $_{\text {Vect }_{k}}(V)$ given by $a \mapsto a \mathrm{id}_{V}$, which by Proposition 8.2.9 implies in the existence of a unique morphism of rings $\Psi: k[x] \rightarrow \operatorname{End}_{\text {Vect }_{k}}(V)$ such that the following diagram commutes

and that $x \stackrel{\Psi}{\mapsto} \phi$. Since $k$ and $x$ generate $k[x]$, for any $f(x):=\sum_{j=1}^{n} a_{j} x^{j} \in k[x]$, we have a mapping

$$
f(x)=\sum_{j=1}^{n} a_{j} x^{j} \stackrel{\Psi}{\longmapsto} a_{0} \operatorname{id}_{V}+a_{1} \phi+\cdots+a_{n-1} \phi^{n-1}+a_{n} \phi^{n}:=f(\phi) .
$$

The ring action $\Psi: k[x] \rightarrow \operatorname{End}_{\text {Vect }_{k}}(V)$ induces the structure of a left- $k[x]$-module on $V$ as

$$
p(x) \cdot v:=p(\phi)(v)
$$

for all $p(x) \in k[x]$ and $v \in V$.
Corollary 8.4.9. Given a field $k$, there exists a bijection

$$
\{k[x] \text {-modules }\} \stackrel{\simeq}{\leftrightharpoons}\left\{(V, \phi): V \in \operatorname{Vect}_{k} \text { and } \phi \in \operatorname{End}_{\operatorname{Vect}_{k}}(V)\right\} .
$$

In other words, $k[x]$-modules are equivalent to $k$-vector spaces together with a uniquely determined $k$-linear endomorphism.

Proof. As constructed in Example 8.4.8, for every $k$-vector space $V$ and $k$-linear endomorphism $\phi: V \rightarrow V$, one has a unique left- $k[x]$-module structure induced on $V$. On the other hand, let $M$ be a left- $k[x]$-module and $\psi: M \rightarrow M$ be the group morphism given by $m \mapsto x m$ for all $m \in M$. Notice that for any $a \in k$ and $m \in M$ we have

$$
\psi(a m)=x(a m)=(x a) m=(a x) m=a(x m)=a \psi(a m)
$$

therefore $\psi$ is $k$-linear. Thus $(M, \phi)$ is the corresponding uniquely defined $k$-vector space together with a $k$-linear endomorphism.

## Category of Modules

Definition 8.4.10 (Morphism of modules). Let $R$ be a ring, and both $M$ and $N$ be abelian groups. Let $M_{R}$ and $N_{R}$ denote right- $R$-modules on $M$ and $N$, while ${ }_{R} M$ and ${ }_{R} N$ denote left- $R$-modules on $M$ and $N$. We define the the following:
(a) A morphism between right- $R$-modules $\phi: M_{R} \rightarrow N_{R}$ is a morphism of abelian groups such that, for all $r \in R$ and $m \in M$ we have

$$
\phi(m r)=\phi(m) r .
$$

(b) A morphism between left-R-modules $\psi:{ }_{R} M \rightarrow{ }_{R} N$ is a morphism of abelian groups such that, for all $r \in R$ and $m \in M$ we have

$$
\psi(r m)=r \psi(m) .
$$

Morphisms of $R$-modules can also be compactly named $R$-linear morphisms.
Definition 8.4.11 (Category of $R$-modules). Given a ring $R$, we denote by $\operatorname{Mod}_{R}$ the category whose objects are right- $R$-modules and morphisms between them. Analogously, we define ${ }_{R}$ Mod to be the category whose objects are left-R-modules and morphisms between them.

If $R$ is a commutative ring, we simply denote the category of modules over $R$ and morphisms between them by $R$-Mod.

Proposition 8.4.12 (Z-modules). Abelian groups are Z-modules in exactly one way. Therefore $\mathrm{Z}_{\mathrm{Mod}}$ and Ab are isomorphic categories.

Proof. Let $G$ be an abelian group. Since $\mathbf{Z}$ is initial in the category of rings and $\operatorname{End}_{\mathrm{Ab}}(G)$ forms a ring, we find that there exists a unique morphism of rings

$$
\mathbf{Z}---\operatorname{End}_{\mathrm{Ab}}(G)
$$

defining an action of $\mathbf{Z}$ on the group $G$.
Example 8.4.13 (Q-vector space). Let $G$ be an abelian group. If there exists a Q-vector space structure on $G$, this structure is unique.

Let $\mu, \sigma: \mathbf{Q} \rightrightarrows \operatorname{End}_{\mathrm{Ab}}(G)$ be two $Q$-module structures on $G$. Since the inclusion $\iota: \mathbf{Z} \hookrightarrow \mathbf{Q}$ is an epimorphism of rings, it follows that, since there exists a unique $\mathbf{Z}$ module structure $\mathbf{Z} \rightarrow \operatorname{End}_{\mathrm{Ab}}(G)$ (see Proposition 8.4.12), it follows that $\mu \iota=\sigma \iota$. On the other hand, since $\iota$ is an epimorphism, then $\mu=\sigma$.

Proposition 8.4.14 (Zero object). Let $R$ be a ring. The trivial group 0 has a unique $R-$ module structure and defines a zero object in the category of (left or right) $R$-modules.

Proof. Notice that the only $R$-module structure on the trivial group 0 is given by $r \cdot 0:=0$ for all $r \in R$. Moreover, for any $R$-module $M$, we have unique $R$-linear morphisms $0 \rightarrow M$ mapping $0 \mapsto 0_{M}$ and $M \rightarrow 0$ mapping $m \mapsto 0$ for all $m \in M$.

Proposition 8.4.15 (Isomorphisms). Let $R$ be a ring. A morphism of (left or right) $R$-modules is an isomorphism if and only if it is a bijective set-function.

Proof. We prove for right- $R$-modules, the proof for left- $R$-modules is completely analogous. Let $\phi: M \rightarrow N$ be an isomorphism of $R$-modules and $\psi: N \rightarrow M$ be its inverse. Since $\phi \psi=\operatorname{id}_{N}$, then $\operatorname{im} \phi=N$, that is, $\phi$ is surjective. On the other hand, since $\psi \phi=\mathrm{id}_{M}$ and $\psi$ is a well defined set-function, it follows that $\phi$ is injective.

Conversely, let $\phi: M \rightarrow N$ be an $R$-linear morphism and a bijective set-function. Let $\psi: N \rightarrow M$ be its inverse as a set-function - we shall prove that $\psi$ is an $R$-linear morphism. Let $n, n^{\prime} \in N$ be any elements, if $\phi(m)=n$ and $\phi\left(m^{\prime}\right)=n^{\prime}$ then $\phi\left(m+m^{\prime}\right)=$
$\phi(m)+\phi\left(m^{\prime}\right)=n+n^{\prime}-$ thus from construction $\psi\left(n+n^{\prime}\right)=m+m^{\prime}=\psi(n)+\psi\left(n^{\prime}\right)$. Now, if $r \in R$ is any ring element, then since $\phi(m r)=\phi(m) r=n r$, we find that $\psi(n r)=m r=\psi(n) r$ - therefore $\psi$ is indeed a morphism of $R$-modules and hence an inverse morphism for $\phi$.

Example 8.4.16. Let $R$ be an integral domain, and $(a) \subseteq R$ be a non-zero principal ideal of $R$. There exists a natural isomorphism of $R$-modules $R \simeq(a)$.

Consider the map $\phi: R \rightarrow(a)$ given by $r \mapsto r a$. Then for any $r, s \in R$ we have

$$
\begin{aligned}
\phi(r+s) & =(r+s) a=r a+s a=\phi(r)+\phi(s) \\
\phi(r s) & =(r s) a=r(s a)=r \phi(s)
\end{aligned}
$$

thus $\phi$ is an $R$-module morphism. The morphism is also clearly surjective by the definition of an ideal. Moreover, one should note that $\phi(1)=a$, which is, by hypothesis, non-zero. Given any two $r, r^{\prime} \in R$ such that $\phi(r)=\phi\left(r^{\prime}\right)$ we obtain

$$
0=\phi(r)-\phi\left(r^{\prime}\right)=r \phi(1)-r^{\prime} \phi(1)=\left(r-r^{\prime}\right) \phi(1)=\left(r-r^{\prime}\right) a,
$$

and since $R$ is an integral domain, it follows that $r=r^{\prime}$ - therefore, $\phi$ is injective. We conclude that $\phi$ is a bijection between $R$-modules and therefore establishes an isomorphism $R \simeq(a)$.

Example 8.4.17 (Morphisms form an $R$-module). Given a ring $R$ and left- $R$-modules $M$ and $N$, the collection of $R$-linear morphisms $\operatorname{Mor}_{R} \operatorname{Mod}(M, N)$ can be endowed with the structure of an right-R-module.

Since $\operatorname{Mor}\left({ }_{R} \operatorname{Mod}\right) \subseteq \operatorname{Mor}(\mathrm{Ab})$, it follows that $\operatorname{Mor}_{R} \operatorname{Mod}(M, N)$ is an abelian group given by

$$
(f+g)(m):=f(m)+g(m)
$$

for any morphisms $f, g: M \rightrightarrows N$ and element $m \in M$. Moreover, one can endow $\operatorname{Mor}_{R} \operatorname{Mod}(M, N)$ with the right ring action $\mathbf{R} \times \operatorname{Mor}_{R} \operatorname{Mod}(M, N) \rightarrow \operatorname{Mor}_{R} \operatorname{Mod}(M, N)$ given by

$$
(f \cdot r)(m):=f(r m)
$$

for every morphism $f: M \rightarrow N$, and any elements $r \in R$ and $m \in M$.
Notice that we've emphasized that, in general, we can only give a right $R$-module structure to $\operatorname{Mor}_{R} \operatorname{Mod}(M, N)$, if on the contrary we defined an left- $R$-action by $(r \cdot f)(m):=$ $f(r m)$, one would suffer from the following problem:

$$
\begin{equation*}
((r s) f)(m)=(r(s f))(m)=(s f)(r m)=f(s(r m))=f((s r) m)=((s r) f)(m) \tag{8.4}
\end{equation*}
$$

where $f: M \rightarrow N$ is a morphism, and $r, s \in R$ and $m \in M$ are any elements. Mind that Eq. (8.4) does not yield a left-R-module structure unless $R$ is commutative.

Lemma 8.4.18. Let $R$ be a ring and let $M, N \in{ }_{R}$ Mod. We have the following properties:
(a) The functor $\operatorname{Mor}_{R} \operatorname{Mod}(M,-):{ }_{R} \operatorname{Mod} \rightarrow \mathrm{Ab}$ is additive.
(b) The collection of morphisms $\operatorname{Mor}_{R} \operatorname{Mod}(M, N)$ has a natural structure of left $Z(R)$ module with multiplication $r f$ mapping $x \mapsto f(r x)$ for any $r \in Z(R)$ and morphism $f: M \rightarrow N$, moreover, addition is defined point-wise as expected. Therefore $\operatorname{Mor}_{R} \operatorname{Mod}(M,-)$ can be seen as an additive functor ${ }_{R} \operatorname{Mod} \rightarrow Z(R) \operatorname{Mod}$.

Proposition 8.4.19 (Additive functor, zero maps \& zero objects). Let $F: R$-Mod $\rightarrow \mathrm{Ab}$ be an additive functor of either variance ${ }^{2}$. Then $F$ maps the zero map $0: M \rightarrow N$ to the zero map $F 0=0: F M \rightarrow F N$, and maps the zero object $\{0\} \in R$-Mod to the zero object $F\{0\}=\{0\} \in \mathrm{Ab}$.

Proof. Since $F$ is additive, for any morphism $f: M \rightarrow N$ we have

$$
F f=F(f+0)=F f+F 0,
$$

therefore it must be the case that $F 0: F M \rightarrow F N$ is the zero morphism in Ab . Moreover, since $L=0 \in R$-Mod if and only if $\mathrm{id}_{L}=0: L \rightarrow L$, then from the fact that if $0 \in \operatorname{End}_{R} \operatorname{Mod}(0)$ is the zero morphism, then

$$
F 0=F \mathrm{id}_{0}=\mathrm{id}_{F 0}
$$

and it follows that $F 0=0$.
Proposition 8.4.20. Let $R$ be a commutative ring. Then there exists a canonical isomorphism of $R$-modules

$$
\operatorname{Mor}_{R-\operatorname{Mod}}(R, M) \simeq M
$$

Proof. Notice that every $R$-module morphism $R \rightarrow M$ has to be of the form $r \mapsto r m$ for some fixed $m \in M$ - for convenience, name this morphism $f_{m}$. We define a map $\phi: \operatorname{Mor}_{R-\operatorname{Mod}}(R, M) \rightarrow M$ by sending $f_{m} \mapsto m$, which is certainly surjective. Moreover, if $f_{m}=f_{n}$ then in particular $m=f_{m}(1)=f_{n}(1)=n$, therefore $\phi$ is injective. Notice that for any $m, n \in M$ we have

$$
f_{m}(r)+f_{n}(r)=r m+r n=r(m+n)=f_{m+n}(r),
$$

thus $f_{m}+f_{n}=f_{m+n}$. Also, given any $s \in R$ we have

$$
s f_{m}(r)=s(r m)=(s r) m=(r s) m=r(s m)=f_{s m}(r)
$$

then $s f_{m}=f_{s m}$. The bijection $\phi$ is also an $R$-module morphism since, given any two $f_{m}, f_{n} \in \operatorname{Mor}_{R-\operatorname{Mod}}(R, M)$ we have

$$
\phi\left(f_{m}+f_{n}\right)=\phi\left(f_{m+n}\right)=m+n=\phi\left(f_{m}\right)+\phi\left(f_{n}\right),
$$

and given $r \in R$,

$$
\phi\left(r f_{m}\right)=\phi\left(f_{r m}\right)=r m=r \phi\left(f_{m}\right) .
$$

Since bijective morphisms are isomorphisms in $R$-Mod, it follows that $\operatorname{Mor}_{R-M o d}(R, M) \simeq$ $M$ via $\phi$.

[^12]Proposition 8.4.21. Let $M$ and $N$ be $R$-modules, for some ring $R$. If $M \simeq N$, then there exists a natural isomorphism between abelian groups

$$
\operatorname{End}_{R-\operatorname{Mod}}(M) \simeq \operatorname{End}_{R-\operatorname{Mod}}(N)
$$

Proof. Let $\phi: M \xrightarrow{\simeq} N$ be an isomorphism. We define a map $\Phi: \operatorname{End}_{R-M o d}(M) \rightarrow$ $\operatorname{End}_{R-\operatorname{Mod}}(N)$ given by the conjugation $f \mapsto \phi f \phi^{-1}$. This uniquely defines an $R$-module morphism for each endomorphism $f: M \rightarrow M$. Notice that, since $\phi$ is a bijection, given any endomorphism $g: N \rightarrow N$ we may define an $R$-morphism $f: M \rightarrow M$ given by $f:=\phi^{-1} g \phi$, so that $\Phi(f)=\phi\left(\phi^{-1} g \phi\right) \phi^{-1}=g$. Therefore $\Phi$ establishes an isomorphism of abelian groups via conjugation.

## Examples on Nilpotency

Lemma 8.4.22 (Nakayama's lemma, a particular case). Let $R$ be a commutative ring, and $a \in R$ be a nilpotent element. Then $M=0$ if and only if $a M=M$.

Proof. If $a=0$ then the statement is true. Suppose that $a$ is non-zero, and let $n \in \mathbf{Z}_{>0}$ be the minimal positive integer such that $a^{n}=0$. If $M=0$ then obviously $a M=a \cdot 0=$ $0=M$. On the converse, if we assume that $a M=M$, let $m \in M$ be any element. Let $m_{1} \in M$ be an element such that $m=a m_{1}$. By induction, let $m_{j} \in M$ be an element such that $m_{j-1}=a m_{j}$, for $1<j \leqslant k$. If we consider the collection $\left(m_{j}\right)_{j=1}^{k}$ we get a chain of equalities

$$
\begin{equation*}
m=a m_{1}=a^{2} m_{2}=\cdots=a^{k} m_{k}=0 . \tag{4}
\end{equation*}
$$

Therefore $M=0$.
Proposition 8.4.23. Let $R$ be a commutative ring and $\mathfrak{a}$ a nilpotent ideal of $R$. Let $\phi: M \rightarrow N$ be an $R$-module morphism. If the induced $R$-module morphism

$$
\bar{\phi}: \frac{M}{\mathfrak{a} M} \longrightarrow \frac{N}{\mathfrak{a} N} \quad \text { mapping } \quad m+\mathfrak{a} M \longmapsto \phi(m)+\mathfrak{a} N
$$

is a surjection, then so is $\phi$.
Proof. Since $\bar{\phi}$ is surjective, it follows that

$$
\bar{\phi}(M / \mathfrak{a} M)=\frac{\phi(M)+\mathfrak{a} N}{\mathfrak{a} N}=\frac{N}{\mathfrak{a} N},
$$

therefore $N=\phi(M)+\mathfrak{a} N$. Suppose that $k \in \mathbf{Z}_{>0}$ is such that $\mathfrak{a}^{k}=0$ (which exists since $\mathfrak{a}$ is a nilpotent ideal). Then via induction we find

$$
\begin{aligned}
N=\phi(M)+\mathfrak{a} N & =\phi(M)+\mathfrak{a}(\phi(M)+\mathfrak{a} N) \\
& =\phi(M)+\mathfrak{a} \phi(M)+\mathfrak{a}^{2} N \\
& =\phi(M)+\mathfrak{a}^{2} N \\
& =\ldots \\
& =\phi(M)+\mathfrak{a}^{k} N \\
& =\phi(M) .
\end{aligned}
$$

Therefore $\phi$ is surjective, since $N=\phi(M)$.

## Kernels \& Cokernels

Lemma 8.4.24. Kernels and cokernels exist in R-Mod.
Proof. Let $\phi: M \rightarrow N$ be any $R$-module and consider $\operatorname{ker} \phi:=\{m \in M: \phi(m)=0\}$, together with the canonical inclusion $\iota: \operatorname{ker} \phi \hookrightarrow M$. Notice that $\phi \iota=0 \iota=0$. We prove that ( $\operatorname{ker} \phi, \iota$ ) is the equalizer of $(\phi, 0)$ : let $P$ be any other $R$-module and $f: P \rightarrow M$ be any $R$-module morphism such that $\phi f=0 f=0$. We define a map $\bar{f}: P \rightarrow \operatorname{ker} \phi$ given by $\bar{f}=f$. Indeed, since $\phi f=0$, then $\operatorname{im} f \subseteq \operatorname{ker} \phi$, and $\bar{f}$ is a well defined and unique $R$-module morphism making the following diagram commute


Thus $(\operatorname{ker} \phi, \iota)$ is indeed the equalizer of $(f, 0)$.
Now we prove that coker $\phi:=N / \operatorname{im} \phi$ together with the natural projection $\pi: N \rightarrow$ coker $\phi$ is the coequalizer of $(\phi, 0)$. Let $C$ be an $R$-module and $g: N \rightarrow C$ be an $R$ module morphism such that $g \phi=g 0=0$. We define a map $\bar{g}$ : coker $\phi \rightarrow C$ to be given by $n+\operatorname{im} \phi \mapsto g(n)$. This map is well defined since $g(n)=0$ for all $n \in \operatorname{im} \phi$. Also, since $g$ is a morphism, $\bar{g}$ is trivially a morphism of $R$-modules too. Moreover, since $\pi$ is an epimorphism, $\bar{g}$ is the unique morphism for which the following diagram commutes


That is, (coker $\phi, \pi)$ is the coequalizer of $(\phi, 0)$.
Proposition 8.4.25 (Properties of kernels \& cokernels). Kernels and cokernels exist in $R$-Mod. Let $\phi: M \rightarrow N$ be an $R$-module morphism, then:
(a) The following propositions are equivalent:

- The $R$-module morphism $\phi$ is a monomorphism.
- The kernel of $\phi$ is trivial, that is, $\operatorname{ker} \phi=0$.
- The set-function induced by $\phi$ is injective.
(b) The following propositions are equivalent:
- The $R$-module morphism $\phi$ is an epimorphism.
- The cokernel of $\phi$ is trivial, that is, coker $\phi=0$.
- The set-function induced by $\phi$ is surjective.

Proof. (a) - Suppose $\phi$ is a monomorphism, and consider $t$ : $\operatorname{ker} \phi \hookrightarrow M$ and 0 , then $\phi \iota=\phi 0=0$, thus $\iota=0-$ which implies in $\operatorname{ker} \phi=0$.

- If we now suppose that ker $\phi=0$, given any two elements $m, m^{\prime} \in M$ such that $\phi(m)=\phi\left(m^{\prime}\right)$ we obtain $0=\phi(m)-\phi\left(m^{\prime}\right)=\phi\left(m-m^{\prime}\right)$ then $m-m^{\prime} \in \operatorname{ker} \phi$, implying in $m=m^{\prime}$.
- Lastly, if $\phi$ is injective, then given any two morphisms $\alpha, \beta: P \rightrightarrows M$ such that $\phi \alpha=\phi \beta$, we have for all $p \in P$ that $\phi(\alpha(p))=\phi(\beta(p))$ - which implies in $\alpha(p)=\beta(p)$, thus $\alpha=\beta$. We conclude that $\phi$ is a monomorphism.
(b) - Suppose that $\phi$ is an epimorphism, then if we consider the canonical inclusiont $\iota: \operatorname{im} \phi \hookrightarrow N$ and the identity morphism $\operatorname{id}_{N}: N \rightarrow N$, one has that $\iota \phi=\mathrm{id}_{N} \phi$. Since $\phi$ is an epimorphism, then $\iota=\operatorname{id}_{N}$ — which is only possible if im $\phi=N$. Therefore coker $\phi$ is trivial.
- If we suppose that coker $\phi$ is trivial, then $\operatorname{im} \phi=N$ and $\phi$ is therefore surjective.
- Suppose that $\phi$ is surjective, and let $\alpha, \beta: N \rightarrow P$ be $R$-module morphisms such that $\alpha \phi=\beta \phi$. Then, for every $n \in N$, there exists $m \in M$ such that $\phi(m)=n$, thus $\alpha(n)=\alpha(\phi(m))=\beta(\phi(m))=\beta(n)$ - therefore $\alpha=\beta$ and $\phi$ is an epimorphism.

Example 8.4.26 (Right \& left inverses). One should not be deceived by the ideas permeating Set, just as in general categories, $R$-Mod monomorphisms and epimorphisms need not have left-inverse and right-inverse, respectively. An example of a monomorphism without a left-inverse is $\mathbf{Z} \mapsto \mathbf{Z}$ given by $a \mapsto 2 a$. On the other hand, the projection $\mathbf{Z} \rightarrow \mathbf{Z} / 2 \mathbf{Z}$ is an epimorphism, although it does not have a right-inverse.

## $R$-Algebras

Example 8.4.27 ( $R$-modules from ring morphisms). Let $\phi: R \rightarrow S$ be any ring morphism. We can induce on $S$ a structure of left- $R$-module by defining a map $\rho: R \times S \rightarrow S$ to map $(r, s) \mapsto \phi(r) s$ for every $r \in R$ and $s \in S$. In particular, this shows that we can endow $R$ itself with a left- $R$-module structure. These constructions can be analogously done for right- $R$-modules as well.

If $R$ happens to be commutative and $\operatorname{im} \phi \subseteq Z(S)$, then by Example 8.4.6 we find that the left and right $R$-module structures on $S$ induced by $\phi$ coincide. Moreover, notice that, given any $s, s^{\prime} \in S$ and $r, r^{\prime} \in R$, we have that

$$
\begin{aligned}
\rho(r, s) \rho\left(r^{\prime}, s^{\prime}\right) & =(\phi(r) s)\left(\phi\left(r^{\prime}\right) s^{\prime}\right)=\phi(r)\left(s \phi\left(r^{\prime}\right)\right) s^{\prime}=\phi(r)\left(\phi\left(r^{\prime}\right) s\right) s^{\prime} \\
& =\left(\phi(r) \phi\left(r^{\prime}\right)\right)\left(s s^{\prime}\right)=\phi\left(r r^{\prime}\right)\left(s s^{\prime}\right) \\
& =\rho\left(r r^{\prime}, s s^{\prime}\right) .
\end{aligned}
$$

This shows that the $R$-module structure induced by $\phi$ is compatible with the ring structure of $S$. This kind of $R$-module receive the name of $R$-algebra, which we now define for later reference.

Definition 8.4.28 ( $R$-algebra). Let $R$ be a commutative ring. We define an $R$-algebra to be a ring morphism $\phi: R \rightarrow S$ such that the image of $\phi$ is contained in the centre of $S$ -
inducing an $R$-module structure on $S$. If the ring $S$ itself is commutative, we say that it has a structure of a commutative $R$-algebra.

We define a morphism $\gamma: \alpha \rightarrow \beta$ of $R$-algebras $\alpha: R \rightarrow S$ and $\beta: R \rightarrow Q$ to be a ring morphism $\gamma: S \rightarrow Q$ such that for all $r \in R$ and $s \in S$ we have $\phi(r s)=r \phi(s)$, and that the following diagram commutes in Ring


We denote by $R$-Alg the category consisting of $R$-algebras and morphisms between them. A particularly important subcategory is that of the commutative $R$-algebras, which we shall denote by $R$-CAlg.

Corollary 8.4.29. Given a commutative ring $R$, then $R$-CAlg is a subcategory of $R /$ CRing — where $R$ /CRing denotes the slice category under $R$.

Example 8.4.30. The category R-CAlg, however, is not a full subcategory of CRing. Notice for instance that the map $\mathbf{C} \rightarrow \mathbf{C}$ given by the complex conjugation $z \mapsto \bar{z}$ is a ring automorphism, nonetheless it isn't a morphism of C-modules.

Proposition 8.4.31 (Initial object). Given a ring $R$, the module structure of $R$ over itself is the initial object of $R-\mathrm{Alg}$.

Proof. Given any $R$-algebra $\phi: R \rightarrow A$ we have, for all $r, r^{\prime} \in R$, that $\phi\left(r r^{\prime}\right)=r \phi\left(r^{\prime}\right)-$ thus $\phi$ is an $R$-algebra morphism. In particular, for $r^{\prime}=1_{R}$ we obtain $\phi(r)=r \phi\left(1_{R}\right)=$ $r 1_{A}=r$ - thus $\phi$ is the unique $R$-algebra morphism $R \rightarrow A$.

Definition 8.4.32 (Field extension). Let $k$ be a field. We define a field extension of $k$ to be a commutative $k$-algebra $K$ with injective ring morphisms inj, inv: $k \rightrightarrows K$ such that $\operatorname{inj}(a) \operatorname{inv}(a)=1$ for all $a \in k$.

Example 8.4.33. Let $k$ be a subfield $k \subseteq \ell$ of a field $\ell$. Then $\ell$ has a natural structure of field extension of $k$. Indeed, we have a natural inclusion $k \hookrightarrow \ell$ mapping $a \mapsto a$ and an inversion map $k \hookrightarrow \ell$ sending $a \mapsto a^{-1}$.

Definition 8.4.34 (Rees algebra). Let $R$ be a commutative ring and $\mathfrak{a}$ ideal of $R$. We define a ring

$$
\operatorname{Rees}_{R}(\mathfrak{a}):=\bigoplus_{j \geqslant 0} \mathfrak{a}^{j}
$$

where $\mathfrak{a}^{0}:=R$, with a multiplication given by

$$
\left(a_{j}\right)_{j \geqslant 0} \cdot\left(b_{j}\right)_{j \geqslant 0}:=\left(\sum_{i+k=j} a_{i} b_{k}\right)_{j \geqslant 0} \in \operatorname{Rees}_{R}(\mathfrak{a}) .
$$

The ring morphism $R \rightarrow \operatorname{Rees}_{R}(\mathfrak{a})$ mapping $r \mapsto(r, 0, \ldots, 0, \ldots)$ is called the Rees algebra of $\mathfrak{a}$.

Proposition 8.4.35. Let $R$ be commutative ring and $a \in R$ be a non-zero-divisor. There exists a natural isomorphism of $R$-algebras

$$
\operatorname{Rees}_{R}((a)) \simeq R[x]
$$

Proof. Notice that $R[x]$ is realized as an $R$-algebra by the inclusion $t: R \hookrightarrow R[x]$ mapping to constant polynomials. Denote, for every $j \geqslant 0$, by $e_{j} \in \operatorname{Rees}_{R}((a))$ the element whose $j$-th coordinate is $a^{j}$ and zero elsewhere. Define a map $\phi: R[x] \rightarrow \operatorname{Rees}_{R}((a))$ by mapping $\phi(r):=r e_{0}$ and $\phi(x):=e_{1}$. Since $\{R, x\}$ generate $R[x]$, this completely defines $\phi$ as a ring morphism, since $\phi\left(r x^{j}\right)=r \phi(x)^{j}=r e_{1}^{j}=r e_{j}$. Moreover, given any $\left(a_{j}\right)_{j \geqslant 0} \in \operatorname{Rees}_{R}((a))$, since $R$ is commutative, we have $a_{j}=r_{j} a^{j}$ for some $r_{j} \in R$. If we let $\left(r_{j}\right)_{j \geqslant 0}$ be the collection of these associated $R$ terms, we can build a polynomial $p(x):=\sum_{j \geqslant 0} r_{j} x^{j}$ so that

$$
\phi(p(x))=\sum_{j \geqslant 0} \phi\left(r_{j} x^{j}\right)=\sum_{j \geqslant 0} r_{j} e_{j}=\left(r_{j} a^{j}\right)_{j \geqslant 0}=\left(a_{j}\right)_{j \geqslant 0} .
$$

Therefore $\phi$ is surjective. Moreover, it is simple to see that if $p(x):=\sum_{j \geqslant 0} b_{j} x^{j}$ and $q(x):=\sum_{j \geqslant 0} c_{j} x^{j}$ are polynomials in $R[x]$ such that $\phi(p(x))=\phi(q(x))$, then their coefficients match - that is, $b_{j}=c_{j}$ for all $j \geqslant 0$, and then $p(x)=q(x)$, making $\phi$ injective. Therefore $\phi$ is an isomorphism of $R$-algebras.

Proposition 8.4.36. Let $R$ be a commutative ring, $a \in R$ be a non-zero-divisor, and $\mathfrak{b} \subseteq R$ be any ideal. Then there exists a natural isomorphism of $R$-algebras

$$
\operatorname{Rees}_{R}(a \mathfrak{b}) \simeq \operatorname{Rees}_{R}(\mathfrak{b})
$$

Proof. Define a map $\phi: \operatorname{Rees}_{R}(\mathfrak{b}) \rightarrow \operatorname{Rees}_{R}(a \mathfrak{b})$ by $\left(b_{j}\right)_{j \geqslant 0} \mapsto\left(a^{j} b_{j}\right)_{j \geqslant 0}-$ since $a$ is a non-zero-divisor, $a^{j} b_{j}=0$ if and only if $b_{j}=0$, therefore $\phi$ is bijective. Notice that $\phi$ satisfies $\phi(r u)=r \phi(u)$ for any $r \in R$ and $u \in \operatorname{Rees}_{R}(\mathfrak{b})$. Moreover for any two $\left(u_{j}\right)_{j \geqslant 0},\left(v_{j}\right)_{j \geqslant 0} \in \operatorname{Rees}_{R}(\mathfrak{b})$, we have

$$
\begin{aligned}
\phi\left(\left(u_{j}\right)_{j} \cdot\left(v_{j}\right)_{j}\right) & =\phi\left(\sum_{i+k=j} u_{i} v_{k}\right)_{j}=\left(a^{j} \sum_{i+k=j} u_{i} v_{k}\right)_{j} \\
& =\left(\sum_{i+k=j} a^{i+k} u_{i} v_{k}\right)_{j}=\left(a^{j} u_{j}\right)_{j} \cdot\left(a^{j} v_{j}\right)_{j} \\
& =\phi\left(\left(u_{j}\right)_{j}\right) \cdot \phi\left(\left(v_{j}\right)_{j}\right)
\end{aligned}
$$

Thus $\phi$ is an $R$-algebra isomorphism between $\operatorname{Rees}_{R}(a \mathfrak{b})$ and $\operatorname{Rees}_{R}(\mathfrak{b})$.

## Submodules \& Quotients

## Submodules

Definition 8.4.37 (Submodule). Let $R$ be a ring and $M$ be an $R$-module. An $R$-module $N$ is said to be a submodule of $R$ if $N \subseteq M$ and the inclusion map $N \hookrightarrow M$ is a morphism of $R$-modules.

Example 8.4.38 (Ideals are the submodules of $R$ ). Let $R$ be a ring endowed with the canonical left- $R$-module structure. The submodules of $R$ are correspond exactly to the left-ideals of $R$.

Example 8.4.39 (Kernel and image are submodules). Given a morphism $\phi: M \rightarrow N$ of $R$-modules, the kernel of $\phi$ is a submodule of $M$, while the image of $\phi$ is a submodule of $N$.

Suppose $M$ and $N$ are right- $R$-modules, for left- $R$-modules the proof is analogous. If $m \in \operatorname{ker} \phi$, then for all $r \in R$ we have $\phi(m r)=\phi(m) r=0 \cdot r=0$ - thus $m r \in \operatorname{ker} \phi$, making $\operatorname{ker} \phi \hookrightarrow M$ a morphism of right-R-modules. Now, if $n \in \operatorname{im} \phi$, there must exist $m \in M$ such that $\phi(m)=n$, therefore for all $r \in R$ we have $\phi(m r)=\phi(m) r=n r \in \operatorname{im} \phi$. We conclude that, indeed, $\operatorname{ker} \phi$ and $\operatorname{im} \phi$ are submodules of $M$ and $N$, respectively.

Example 8.4.40 (Intersection \& sum of submodules). Let $M$ be an $R$-module and $\left(N_{j}\right)_{j \in J}$ be a collection of submodules of $M$. We have the following:
(a) The intersection $\bigcap_{j \in J} N_{j}$ is a submodule of $M$.
(b) The sum

$$
\sum_{j \in J} N_{j}:=\left\{\sum_{j \in F} n_{j}: F \subseteq J \text { is finite, and } n_{j} \in N_{j} \text { for all } j \in F\right\}
$$

is a submodule of $M$.
Proof. For the intersection, given any two $a, b \in \bigcap_{j \in J} N_{j}$ we have, for each $j \in J$, that $a, b \in N_{j}$ and therefore $a+b \in N_{j}$ - this implies in $a+b \in \bigcap_{j \in J} N_{j}$. Moreover, given any element $r \in R$ it is equally clear that ar $\in N_{j}$ for every $j \in J$, which implies in ar $\in \bigcap_{j \in J} N_{j}$. Therefore $\bigcap_{j \in J} N_{j}$ is indeed a submodule of $M$.

For the ease of notation, define $N:=\sum_{j \in J} N_{j}$. Notice that since every sum in $N$ is finite, then $N \subseteq M$. Given any two elements $\sum_{j \in F} a_{j}, \sum_{j \in F^{\prime}} b_{j} \in N$, we have that $F \cup F^{\prime} \subseteq J$ is finite, hence we have an element $\sum_{j \in F \cup F^{\prime}} m_{j} \in N$ given by

$$
m_{j}:= \begin{cases}a_{j}, & j \in F \backslash F^{\prime} \\ b_{j}, & j \in F^{\prime} \backslash F \\ a_{j}+b_{j}, & j \in F \cap F^{\prime}\end{cases}
$$

It is easy to see that this constructed element satisfies

$$
\sum_{j \in F \cup F^{\prime}} m_{j}=\left(\sum_{j \in F} a_{j}\right)+\left(\sum_{j \in F^{\prime}} b_{j}\right)
$$

therefore $N$ is closed under finite addition. Since multiplication is distributive over finite sums, it follows that for any $r \in R$ we have

$$
\left(\sum_{j \in F} a_{j}\right) r=\sum_{j \in F} a_{j} r
$$

and since $a_{j} r \in N_{j}$ for each $j \in F$, it follows that $\left(\sum_{j \in F} a_{j}\right) r \in N$. Thus $N$ is a submodule of $M$.

Example 8.4.41 (Union of submodules). Let $M$ be an $R$-module, for a ring $R$. The following are two propositions concerning submodules of $M$ :

1. Given submodules $S, T \subseteq M$, the union $S \cup T$ is a submodule of $M$ if and only if $S \subseteq T$ or $T \subseteq S$.
2. Let $\left(N_{j}\right)_{j \in \mathbf{N}}$ be an ascending chain of submodules of $M$-that is, $N_{j} \subseteq N_{j+1}$ for all $j \in \mathbf{N}$. Then the union $\bigcup_{j \in \mathbf{N}} N_{j}$ is a submodule of $M$.

Proof. For item (a), if $S \cup T$ is a submodule of $M$, let $s \in S$ and $t \in T$ be any two elements, then $s+t \in S \cup T$. This implies that either $s+t \in S$ (which would imply in $t \in S$ ) or $s+t \in T$ (which would imply in $s \in T$ ), therefore either $S \subseteq T$ or $T \subseteq S$. For the converse, suppose, without loss of generality, that $S \subseteq T$, then $S \cup T=S$, therefore $S \cup T$ is a submodule of $M$.

We now prove item (b). For the sake of notation, let $N:=\bigcup_{j \in \mathbf{N}} N_{j}$. If $n, n^{\prime} \in N$ are any two elements, there must exist indices $j, j^{\prime} \in \mathbf{N}$ such that $n \in N_{i}$ and $n^{\prime} \in N_{i^{\prime}}$ for all $i>j$ and for all $i^{\prime}>j^{\prime}$. Define $k:=\max \left(j, j^{\prime}\right)$, then $n, n^{\prime} \in N_{i}$ for all $i>k$-implying in $n+n^{\prime} \in N_{i}$ and $r n \in N_{i}$ for any $r \in R$. Therefore $n+n^{\prime} \in N$ and $r n \in N$, hence $N$ is a submodule of $M$.

Example 8.4.42 (Totally ordered submodules). Let $M$ be a finite Z-module such that the collection of its submodules is totally ordered with respect to inclusion. Then there exists a prime $p$ such that the number of elements of $M$ is a power of $p$.

Proof. Let $|M|:=d$ and suppose there exists two primes $p$ and $q$ dividing $d$. By Proposition 7.4.29 we know that there must exist $m, n \in M$ with order $p$ and $q$, respectively. We now consider the submodules $\langle m\rangle$ and $\langle n\rangle$ of $M$. Since $M$ has a totally ordered set of submodules, we may assume without loss of generality that $\langle m\rangle \subseteq\langle n\rangle$ —therefore there exists $a \in \mathbf{Z}$ such that $m=a n$. Notice that $p m=p a n=0$, therefore $p a$ must be a divisor of $q$-but since $q$ is prime, $p a$ is either 1 or $q$. Since $p$ is also prime, it must be the case that $a=1$ and $p=q$. Thus there exists a unique prime divisor of $d$-hence $d=p^{\alpha}$ for some $\alpha \in \mathbf{Z}$.

## Simple $R$-Modules

Definition 8.4.43 (Simple module). Let $R$ be a ring. An $R$-module $M$ is said to be simple if its only submodules are $\{0\}$ and $M$ itself.

Lemma 8.4.44 (Schur's). Let $M$ and $N$ be simple $R$-modules. If $\phi: M \rightarrow N$ is an $R$-module morphism, then either $\phi=0$ or $\phi$ is an isomorphism.

Proof. Since ker $\phi$ is a submodule of $M$, it can either be 0 or $M$. If $\operatorname{ker} \phi=M$, then $\phi=0$. On the other hand, if $\operatorname{ker} \phi=0$ then $\phi$ is an injective morphism. Moreover, since $\operatorname{im} \phi$ is a submodule of $N$, it's either 0 or $N$ - since ker $\phi$ in trivial, then the only possibility is that $\operatorname{im} \phi=N$, thus $\phi$ is surjective.

Corollary 8.4.45. Let $M$ and $N$ be right- $R$-modules and consider $\operatorname{Mor}_{M_{M o d}^{R}}(M, N)$ as a ring. If $M$ is simple, then $\operatorname{Mor}_{M_{R o d}^{R}}(M, N)$ is a division ring.

Proof. Let $\mathfrak{a}$ be a left-ideal (or right-ideal) of $\operatorname{Mor}_{M_{M o d_{R}}}(M, N)$. If there exists an isomorphism $\phi \in \mathfrak{a}$, then $\phi^{-1} \phi=\operatorname{id}_{M} \in \mathfrak{a}$ and therefore, for all elements $\psi \in \operatorname{Mor}_{M_{M o d_{R}}}(M, N)$ we have $\psi \operatorname{id}_{M}=\psi \in \mathfrak{a}$ - thus $\mathfrak{a}=\operatorname{Mor}_{\operatorname{Mod}_{R}}(M, N)$ (the same equivalent argument can be used for right-ideals, where instead we get $\phi \phi^{-1}=\mathrm{id}_{N}$ in the ideal). If there exist no isomorphism in $\mathfrak{a}$, by Lemma 8.4.44 it is only composed of the zero morphism, thus $\mathfrak{a}=0$. From Proposition 8.3.14 we conclude that $\operatorname{Mor}_{\text {Mod }_{R}}(M, N)$ is a division ring.

Example 8.4.46. We now show a counterexample ilustrating why the opposite of Corollary 8.4.45 does not hold in general. Let $k$ be a field and

$$
A:=\left[\begin{array}{ll}
1 & 1 \\
0 & 0
\end{array}\right] \quad B:=\left[\begin{array}{ll}
0 & 0 \\
0 & 1
\end{array}\right]
$$

be matrices. Define $R$ to be the $k$-algebra generated by $A$ and $B$. Consider the left- $R-$ module $M=k^{2}$, with left-multiplication by matrices. We'll show that $\operatorname{Mor}_{R} \operatorname{Mod}(M, M)$ is a division ring, while $M$ is not simple.

Consider the principal ideal $\mathfrak{a}:=\left(\left[\begin{array}{l}1 \\ 0\end{array}\right]\right)$ of $M$. Notice that for any $a, b \in k$ we have

$$
(a A+b B)\left[\begin{array}{l}
1 \\
0
\end{array}\right]=\left[\begin{array}{ll}
a & a \\
0 & b
\end{array}\right]\left[\begin{array}{l}
1 \\
0
\end{array}\right]=\left[\begin{array}{l}
a \\
0
\end{array}\right]
$$

thus $\mathfrak{a} \neq M$ and $\mathfrak{a} \neq 0$, and $M$ isn't simple.
Let $\mathfrak{h}$ be a left-ideal (or right-ideal) of the ring $\operatorname{Mor}_{R} \operatorname{Mod}(M, M)$. If $\mathfrak{h} \neq 0$, consider a non-zero morphism $M=\left[\begin{array}{ccc}a & b \\ c & d\end{array}\right] \in \mathfrak{h}$. Since $M$ is a left- $R$-module morphism, it must be the case that, for any $x=\left[\begin{array}{l}x_{1} \\ x_{2}\end{array}\right] \in M$, we have $M(A x)=A M(x)$ and $M(B x)=B M(x)$, but

$$
\begin{align*}
& M(A x)=\left[\begin{array}{l}
a\left(x_{1}+x_{2}\right) \\
c\left(x_{1}+x_{2}\right)
\end{array}\right]  \tag{8.5}\\
& M(B x)=\left[\begin{array}{l}
b x_{2} \\
d x_{2}
\end{array}\right] \tag{8.6}
\end{align*}
$$

while on the other hand we have

$$
\begin{align*}
& A M(x)=\left[\begin{array}{c}
(a+c) x_{1}+(b+d) x_{2} \\
0
\end{array}\right]  \tag{8.7}\\
& B M(x)=\left[\begin{array}{c}
0 \\
c x_{1}+d x_{2}
\end{array}\right] \tag{8.8}
\end{align*}
$$

Since Eq. (8.5) equals Eq. (8.7), while Eq. (8.6) equals Eq. (8.8) - and since $M$ is nonzero by hypothesis - we obtain a solution $a=d=1$ and $b=c=0$, yielding $M=\left[\begin{array}{ll}1 & 0 \\ 0 & 1\end{array}\right]$, which implies in $\mathfrak{h}=\operatorname{Mor}_{R} \operatorname{Mod}(M, M)$.

Proposition 8.4.47. Let $M$ be an $R$-module. Then $M$ is simple if and only if $M \simeq R / \mathrm{m}$ for a maximal ideal $\mathfrak{m}$ of $R$.

Proof. Suppose $\phi: M \xrightarrow{\simeq} R / \mathrm{m}$ is an isomorphism of $R$-modules. If $N \subsetneq M$ is a proper submodule, then $\phi(N) \subseteq R / \mathfrak{m}$ must also be a submodule of $R / \mathfrak{m}$-that is, an ideal. Since $\mathfrak{m}$ is maximal, then the only ideals of the field $R / \mathfrak{m}$ are either the field itself or the zero ideal $\mathfrak{m}$. Since $\phi$ is an isomorphism and $N$ is proper, it must be the case that $\phi(N)=\mathfrak{m}$. But since $\phi(0)=\mathfrak{m}$, then $N=\{0\}$-thus $M$ is simple.

For the converse, assume that $M$ is simple, so that every non-zero element of $M$ generates the whole module. In particular $1 \in M$ generates $M$, therefore the morphism of $R$-modules $\psi: R \rightarrow M$ mapping $a \mapsto a \cdot 1$ is surjective, therefore

$$
M \simeq R / \operatorname{ker} \psi
$$

By Corollary 8.4.55 if $\mathfrak{a} \subsetneq R$ is any proper ideal (submodule) containing ker $\psi$, then its corresponding submodule is $\mathfrak{a} / \operatorname{ker} \psi \subseteq R / \operatorname{ker} \psi \simeq M$. Since $M$ is simple and $\mathfrak{a}$ is proper, it follows that $\mathfrak{a} / \operatorname{ker} \psi$ must be the zero ideal, therefore $\mathfrak{a}=\operatorname{ker} \psi$-that is $\operatorname{ker} \psi$ is maximal.

## Quotient modules

Definition 8.4.48 (Quotient module). Let $R$ be a ring, and $M$ be a left- $R$-module. If $N \subseteq M$ is a submodule, then in particular $N$ is a normal subgroup of $M$ and therefore $M / N$ is an abelian group. We endow the group $M / N$ with a left- $R$-action $R \times(M / N) \rightarrow M / N$ given by

$$
r(m+N):=r m+N
$$

for all $r \in R$ and $m \in M$. This action turns $M / N$ into a left- $R$-module and the canonical projection $\pi: M \rightarrow M / N$ into an $R$-module morphism. Therefore, the submodule $N$ is the kernel of the canonical projection $\pi$.

Example 8.4.49. Let $R$ be a non-commutative ring and $\mathfrak{a}$ a left-submodule, then the set of equivalence classes $R / \mathfrak{a}$ is not a ring - since $\mathfrak{a}$ is be required to be a two-sided-ideal for the quotient to be a ring. Although not a ring, $R / \mathfrak{a}$ is is an abelian group and one can endow $R / \mathfrak{a}$ with the canonical left- $R$-action given by $r(a+\mathfrak{a})=r a+\mathfrak{a}$ for every $r \in R$ and $a+\mathfrak{a} \in R / \mathfrak{a}$.

Theorem 8.4.50 (Universal property of quotient modules). Let $R$ be a ring and $N$ be a submodule of an $R$-module $M$. For every module $Z$ together with a morphism $\psi: M \rightarrow Z$ of $R$-modules such that $N \subseteq$ ker $\psi$, there exists a unique morphism of rings $\phi: M / N \rightarrow Z$ for which the following diagram commutes


Proof. In particular, the existence and uniqueness of $\phi$ as a group morphism is shown in Proposition 7.4.16 - we only show that $\phi$ is a morphism of $R$-modules. Let's assume
that we are working with right- $R$-modules, the same analogous proof would work for the left modules. Let $m+N \in M / N$ and $r \in R$ be any elements, then $\psi(m r)=\psi(m) r$ and since $\psi=\phi \pi$ it follows that

$$
\phi((m+N) r)=\phi(m r+N)=\psi(m r)=\psi(m) r=\phi(m+N) r
$$

which shows that $\phi$ is indeed a morphism of right- $R$-modules.
Theorem 8.4.51 (Factorization of morphisms). Every R-module morphism $\phi: M \rightarrow N$ factors as follows


Proof. From the universal property Theorem 8.4.50 we find that $\phi$ induces a unique morphism $\bar{\phi}: M / \operatorname{ker} \phi \rightarrow N$ - being defined as $\bar{\phi}(a+\operatorname{ker} \phi)=\phi(a)$ for any $a+$ $\operatorname{ker} \phi \in M / \operatorname{ker} \phi$. We simply restrict the codomain of $\bar{\phi}$ to $\operatorname{im} \phi$ so that it becomes surjective. Moreover, given any two classes $a+\operatorname{ker} \phi, b+\operatorname{ker} \phi \in M / \operatorname{ker} \phi$, such that $\bar{\phi}(a+\operatorname{ker} \phi)=\bar{\phi}(b+\operatorname{ker} \phi)$ then $\phi(a)=\phi(b)$ and thus $a-b \in \operatorname{ker} \phi$ - therefore $a+\operatorname{ker} \phi=b+\operatorname{ker} \phi$ and $\bar{\phi}$ is injective.

Corollary 8.4.52 (First isomorphism). Let $\phi: M \rightarrow N$ be a surjective $R$-module morphism. There exists a canonical isomorphism of $R$-modules

$$
N \simeq M / \operatorname{ker} \phi
$$

Proof. Since $\phi$ is surjective, $\operatorname{im} \phi=N$ and from Theorem 8.4.51 we conclude that the induced morphism $\bar{\phi}: M / \operatorname{ker} \phi \xrightarrow{\simeq} N$ establishes the wanted canonical isomorphism.

Proposition 8.4.53 (Second isomorphism). Let $M$ be an $R$-module and $P, N \subseteq M$ be submodules. There exists a canonical isomorphism

$$
\frac{N+P}{N} \simeq \frac{P}{N \cap P}
$$

Proof. Since $N \hookrightarrow N+P$ mapping $n \mapsto n+0=n$ is clearly an $R$-module morphism, $N$ is a submodule of $N+P$. Moreover, the inclusion $N \cap P \hookrightarrow P$ mapping $p \mapsto p$ is also a morphism since $p \in P$ for all $p \in N \cap P$.

Consider the map $\phi: N+P \rightarrow \frac{P}{N \cap P}$ given by $n+p \mapsto p+N \cap P$, which is clearly an $R$-module morphism. Given any class $p+N \cap P \in \frac{P}{N \cap P}$, one can choose a representative $p \in P$ so that $\phi(p)=p+N \cap P$ - thus $\phi$ is surjective. Moreover, $\phi(n+p)=N \cap P$ if, and only if $p \in N \cap P$, which in this case implies in $n+p \in N$. Therefore ker $\phi=N$. By Corollary 8.4.52 we obtain $\frac{N+P}{\operatorname{ker} \phi} \simeq \frac{P}{N \cap P}$, which is the required isomorphism.

Proposition 8.4.54 (Third isomorphism). Let $M$ be an $R$-module and $N \subseteq M$ be a submodule. If $P \subseteq M$ is a submodule containing $N$ - then $P / N$ is a submodule of $M / N$, and there exists a canonical isomorphism

$$
\frac{M / N}{P / N} \simeq M / P
$$

Proof. Notice that since $N \hookrightarrow P$ is clearly an $R$-module morphism, thus $N$ is a submodule of $P$. Moreover, if we consider the inclusion $i: P / N \hookrightarrow M / N$ by mapping $p+N \mapsto p+N$, since $P \subseteq M$ this is well defined and is a morphism of groups furthermore, given any $r \in R$, one has

$$
i((p+N) r)=i(p r+N)=p r+N=(p+N) r=i(p+N) r
$$

for any $p+N \in P / N$, thus $i$ is an $R$-module morphism. From the last inclusion we conclude that $P / N$ is a submodule of $M / N$.

Let $\phi: M / N \rightarrow M / P$ be defined by mapping $m+N \mapsto m+P-$ since $N \subseteq P$, this map is well defined and and surjective. Moreover, for any $p+N \in M / N$ where $p \in P$, we have $\phi(p+N)=p+P=P$. Moreover, if $m+N \in \operatorname{ker} \phi$ then necessarily $m \in P$. Therefore ker $\phi=P / N$ and by Corollary 8.4.52 $\phi$ induces an isomorphism $\frac{M / N}{\operatorname{ker} \phi} \simeq M / P$.

Corollary 8.4.55 (Submodule correspondence). Let $M$ be an $R$-module and $N \subseteq M$ be a submodule. There exists a bijection

$$
\{\text { submodule } P \text { of } M \text { containing } N\} \longrightarrow\{\text { submodules of } M / N\}
$$

$$
P \longmapsto P / N
$$

Moreover, given submodules $P, P^{\prime} \subseteq M$, we have that $N \subseteq P \subseteq P^{\prime}$ if and only if $P / N \subseteq P^{\prime} / N$ in $M / N$-that is, the bijection preserves inclusions.

Proposition 8.4.56. Let $R$ be a commutative ring, and $\mathfrak{a}, \mathfrak{b} \subseteq R$ be ideals. Then, there exists a natural $R$-module isomorphism

$$
\mathfrak{a} \cdot \frac{R}{\mathfrak{b}} \simeq \frac{\mathfrak{a}+\mathfrak{b}}{\mathfrak{b}}
$$

Proof. Define a map $\phi: \mathfrak{a}+\mathfrak{b} \rightarrow \mathfrak{a} \cdot \frac{R}{\mathfrak{b}}$ given by $\phi(a+b)=a+\mathfrak{b}$ for every $a \in \mathfrak{a}$ and $b \in \mathrm{~b}$. Clearly such a map establishes an $R$-module morphism. Moreover, if $p:=\sum_{j=1}^{n} a_{j} r_{j}+\mathfrak{b} \in \mathfrak{a} \cdot \frac{R}{\mathfrak{b}}$ is any element, since $R$ is commutative then $\mathfrak{a}$ is a two-sided ideal, hence one can choose, for every $1 \leqslant j \leqslant n$, an element $a_{j}^{\prime} \in \mathfrak{a}$ such that $a_{j} r_{j}=a_{j}^{\prime}$. Therefore if we consider $q:=\sum_{j=1}^{n} a_{j}^{\prime} \in \mathfrak{a}+\mathfrak{b}$, one has $\phi(q)=p$ - thus $\phi$ is surjective. Furthermore, an element $a+b \in \mathfrak{a}+\mathfrak{b}$ belongs to ker $\phi$ if and only if $a+b \in \mathfrak{b}-$ therefore $\operatorname{ker} \phi=\mathfrak{b}$. By Corollary 8.4.52 we obtain an isomorphism $\frac{\mathfrak{a}+\mathfrak{b}}{\operatorname{ker} \phi} \simeq \mathfrak{a} \cdot \frac{R}{\mathfrak{b}}$ as wanted.

## Products \& Coproducts

Definition 8.4.57 (Direct product). Let $R$ be a ring and $\left(M_{j}\right)_{j \in J}$ be a collection of $R$ modules (either right or left modules). We define the direct product of $\left(M_{j}\right)_{j \in J}$ as the set

$$
\prod_{j \in J} M_{j}:=\left\{\left(m_{j}\right)_{j \in J}: m_{j} \in M_{j}\right\}
$$

together with coordinate-wise addition and multiplication by elements of $R$. This structure comes naturally with canonical projections $\pi_{i}: \prod_{j \in J} M_{j} \rightarrow M_{i}$ mapping $\left(m_{j}\right)_{j \in J} \mapsto m_{i}$ for all $i \in J$.

Definition 8.4.58 (Direct sum). Let $R$ be a ring and $\left(M_{j}\right)_{j \in J}$ be a collection of $R$-modules (either right or left modules). We define the direct product of this family to be a module

$$
\bigoplus_{j \in J} M_{j}:=\left\{\left(m_{j}\right): m_{j} \in M_{j} \text { and } m_{j} \neq 0 \text { for finitely many } j \in J\right\}
$$

with a natural component-wise addition, and multiplication by elements of $R$. Such a structure also naturally induces canonical inclusions $\iota_{i}: M_{i} \hookrightarrow \bigoplus_{j \in J} M_{j}$ given by $m \mapsto\left(m_{j}\right)_{j \in J}$ where $m_{i}=m$ and $m_{j}=0$ for $j \neq i$.

Theorem 8.4.59. In the category of modules over a given ring, direct products are products and direct sums are coproducts.
Proof. Let $\left(M_{j}\right)_{j \in J}$ be a collection of $R$-modules (either left or right).

- (Product) Define $M:=\prod_{j \in J} M_{j}$. Given an $R$-module $N$ and a family of morphisms $\left(\phi_{j}: N \rightarrow M_{j}\right)_{j \in J}$, we define a map $\phi: N \rightarrow M$ to be given by $n \mapsto$ $\left(\phi_{j}(n)\right)_{j \in J}$. Notice that since each $\phi_{j}$ is a morphism, then given any $n, n^{\prime} \in N$ we have

$$
\begin{aligned}
\phi\left(n+n^{\prime}\right) & =\left(\phi_{j}\left(n+n^{\prime}\right)\right)_{j \in J}=\left(\phi_{j}(n)+\phi_{j}\left(n^{\prime}\right)\right)_{j \in J}=\left(\phi_{j}(n)\right)_{j \in J}+\left(\phi_{j}\left(n^{\prime}\right)\right)_{j \in J} \\
& =\phi(n)+\phi\left(n^{\prime}\right)
\end{aligned}
$$

moreover, if $r \in R$ is any element then

$$
\phi(r n)=\left(\phi_{j}(r n)\right)_{j \in J}=\left(r \phi_{j}(n)\right)_{j \in J}=r\left(\phi_{j}(n)\right)_{j \in J}=r \phi(n) .
$$

That is, $\phi$ is an $R$-module morphism and clearly $\pi_{j} \phi=\phi_{j}$ for all $j \in J$. Moreover, uniqueness comes from the fact that the natural projections are epimorphisms.

- (Coproduct) Define $M:=\bigoplus_{j \in J} M_{j}$. Given an $R$-module $N$ and a family of morphisms $\left(\psi_{j}: M_{j} \rightarrow N\right)_{j \in J}$, we define a map $\psi: M \rightarrow N$ by $\left(m_{j}\right)_{j \in J} \mapsto \sum_{j \in J} \psi_{j}\left(m_{j}\right)-$ which is well defined since $m_{j} \neq 0$ only for finitely many $j \in J$ and thus $\sum_{j \in J} \psi_{j}\left(m_{j}\right)$ constitutes only of finitely many terms, since $\psi_{j}(0)=0$ for any $j \in J$. This map clearly defines a morphism of $R$-modules, and since $\iota_{j}$ are monomorphisms, $\psi$ is the unique morphism such that $\psi \iota_{j}=\psi_{j}$.

Proposition 8.4.60. Let $\left(M_{j}\right)_{j \in J}$ be a collection of $R$-modules and $\left(N_{j}\right)_{j \in J}$ be a corresponding collection of submodules $N_{j} \subseteq M_{j}$. Then $\bigoplus_{j \in J} N_{j}$ is naturally identified as a submodule of $\bigoplus_{j \in J} M_{j}$, and there exists a natural isomorphism

$$
\frac{\bigoplus_{j \in J} M_{j}}{\bigoplus_{j \in J} N_{j}} \simeq \bigoplus_{j \in J} M_{j} / N_{j} .
$$

Proof. For the sake of notation, define $M:=\bigoplus_{j \in J} M_{j}$ and $N:=\bigoplus_{j \in J} N_{j}$. Since the inclusion $N \hookrightarrow M$ is an $R$-module morphism, it follows that $N$ is a submodule of $M$. Consider the natural projections $\left(\pi_{j}: M_{j} \rightarrow M_{j} / N_{j}\right)_{j \in J}$, with kernels ker $\pi_{j}=N_{j}$. Notice that the map $\pi: M \rightarrow \bigoplus_{j \in J} M_{j} / N_{j}$ given by $m \mapsto\left(\pi_{j}(m)\right)_{j \in J}$ has a kernel ker $\pi=N$, and is naturally surjective from its construction. Therefore by the universal property of quotients we find $M / \operatorname{ker} \pi=M / N \simeq \bigoplus_{j \in J} M_{j} / N_{j}$.

Proposition 8.4.61 (Internal sum). Given a ring $R$, let $M$ be an $R$-module (either right or left module) and $\left(N_{j}\right)_{j \in J}$ be a collection of submodules of $M$. The following propositions are equivalent:
(a) The map $\bigoplus_{j \in J} N_{j} \rightarrow \sum_{j \in J} N_{j}$ given by $x \mapsto \sum_{j \in J} \pi_{j}(x)$ is a unique isomorphism of $R$-modules.
(b) For every $i \in J$, we have $N_{i} \cap \sum_{j \in J \backslash i} N_{j}=0$.
(c) If $\sum_{j \in J} x_{j} \in \sum_{j \in J} N_{j}$ is zero, then $x_{j}=0$ for all $j \in J$.
(d) Every element $x \in \sum_{j \in J} N_{j}$ can be uniquely expressed as a sum $x=\sum_{j \in J} x_{j}$ for $x_{j} \in N_{j}$ and $x_{j} \neq 0$ for only finitely many $j \in J$.

Proof. - (a) $\Rightarrow$ (b). Let $n_{i} \in N_{i} \cap \sum_{j \in J \backslash i} N_{j}$ be any element, then one can express $n_{i}$ as $n_{i}=\sum_{j \in J \backslash i} a_{j} n_{j}$ for, therefore $n_{i}-\sum_{j \in J \backslash i} a_{j} n_{j}=0$. By (a) we find that such an element has preimage $0 \in \bigoplus_{j \in J} N_{j}$, which can only be the case if $n_{i}=0$.

- (b) $\Rightarrow$ (c). If $\sum_{j \in J} x_{j}=0$, then for every $i \in J$ we have $x_{i}=-\sum_{j \in J \backslash i} x_{j}$ - but assuming (b) we conclude that $x_{i}=0$.
- $(c) \Rightarrow(d)$. The condition of finitely many non-zero terms is already satisfied from the definition of $\sum_{j \in J} N_{j}$. Now, if $x=\sum_{j \in J} x_{j}=\sum_{j \in J} y_{j}$ are two representations of the same element $x \in \sum_{j \in J} N_{j}$, then $\sum_{j \in J}\left(x_{j}-y_{j}\right)=0$. By (c) this implies in $x_{j}=y_{j}$ for all $j \in J$.
- $(\mathrm{d}) \Rightarrow(\mathrm{a})$. Unicity of expression implies that the given map is injective and unique. Moreover, the map is clearly surjective - therefore it establishes a unique isomorphism.

Proposition 8.4.62. Let $R$ be a ring, and consider a right- $R$-module $M$, and a collection $\left(N_{j}\right)_{j \in J}$ of right-R-modules. The following propositions hold:
(a) There exists a canonical isomorphism of abelian groups

$$
\operatorname{Mor}_{\operatorname{Mod}_{R}}\left(\bigoplus_{j \in J} N_{j}, M\right) \simeq \prod_{j \in J} \operatorname{Mor}_{\operatorname{Mod}_{R}}\left(N_{j}, M\right)
$$

(b) There exists a canonical isomorphism of abelian groups

$$
\operatorname{Mor}_{\operatorname{Mod}_{R}}\left(M, \prod_{j \in J} N_{j}\right) \simeq \prod_{j \in J} \operatorname{Mor}_{\operatorname{Mod}_{R}}\left(M, N_{j}\right)
$$

(c) If $\left(M_{i}\right)_{i=1}^{n}$ and $\left(N_{j}\right)_{j=1}^{m}$ are finite collections of right- $R$-modules, then there exists a canonical isomorphism of abelian groups

$$
\operatorname{Mor}_{\operatorname{Mod}_{R}}\left(\bigoplus_{i=1}^{n} M_{i}, \bigoplus_{j=1}^{m} N_{j}\right) \simeq \bigoplus_{i=1}^{n} \bigoplus_{j=1}^{m} \operatorname{Mor}_{\operatorname{Mod}_{R}}\left(M_{i}, N_{j}\right) .
$$

Proof. (a) Define a map $\Phi: \operatorname{Mor}_{\text {Mod }_{R}}\left(\bigoplus_{j \in J} N_{j}, M\right) \rightarrow \prod_{j \in J} \operatorname{Mor}_{\text {Mod }_{R}}\left(N_{j}, M\right)$ by sending $f \mapsto\left(f \iota_{j}\right)_{j \in J}$, where $\iota_{j}: N_{j} \hookrightarrow \bigoplus_{j \in J} N_{j}$ is the canonical inclusion. Notice that $\Phi$ is indeed a group morphism. Also, morphisms with equal image must agree on every element $\iota_{j}\left(n_{j}\right)$ for any $j \in J$ and $n_{j} \in N_{j}$, therefore, since these elements generate $\bigoplus_{j \in J} N_{j}$, it follows that the morphisms need to be equal - thus $\Phi$ is injective. Moreover, if $\left(g_{j}: N_{j} \rightarrow M\right)_{j \in J}$ is any collection of $R$-linear maps, then by the coproduct universal property there exists a unique $f: \bigoplus_{j \in J} N_{j} \rightarrow M$ such that $f \iota_{j}=g_{j}$ for all $j \in J$.
(b) Define a map $\Psi: \operatorname{Mor}_{\operatorname{Mod}_{R}}\left(M, \prod_{j \in J} N_{j}\right) \rightarrow \prod_{j \in J} \operatorname{Mor}_{\operatorname{Mod}_{R}}\left(M, N_{j}\right)$ by $f \mapsto\left(\pi_{j} f\right)-$ again this product preserves additive structure. Also, this map is injective since, given $f, g: M \rightrightarrows \prod_{j \in J} N_{j}$ such that $\Psi(f)=\Psi(g)$, we must have $\pi_{j} f(m)=\pi_{j} g(m)$ for every $j \in J$ and $m \in M$, therefore $f(m)=g(m)$ and $f=g$. Surjectivity comes from the product universal property: given a collection $\left(g_{j}: M \rightarrow N_{j}\right)_{j \in J}$ of morphisms, there exists a unique morphism $f: M \rightarrow \prod_{j \in J} N_{j}$ such that $\pi_{j} f=g_{j}$ for all $j \in J$.
(c) For finite indexing sets direct sums and products are isomorphic, thus the isomorphism follows directly from the last two items.

Example 8.4.63. Let $M_{1}, \ldots, M_{n}$ be right- $R$-modules and define a right- $R$-module $M:=\bigoplus_{j=1}^{n} M_{j}$ and a ring

$$
H:=\left\{\left[\begin{array}{ccc}
f_{11} & \ldots & f_{1 n} \\
\vdots & \ddots & \vdots \\
f_{n 1} & \ldots & f_{n n}
\end{array}\right]: f_{i j} \in \operatorname{Mor}_{\operatorname{Mod}_{R}}\left(M_{j}, M_{i}\right)\right\} .
$$

Then there exists a natural isomorphism of rings

$$
H \simeq \operatorname{End}_{\operatorname{Mod}_{R}}(M)
$$

First we show that $H$ is indeed a ring. Given any $\left[f_{i j}\right],\left[g_{i j}\right] \in H$, we have entrywise $f_{i j}+g_{i j} \in \operatorname{Mor}_{\text {Mod }_{R}}\left(M_{j}, M_{i}\right)$, thus $\left[f_{i j}+g_{i j}\right] \in H$, and $H$ is an abelian group via matrix addition. Now, if we define a product on $H$ as the matrix product $\left[f_{i j}\right] \cdot\left[g_{i j}\right]:=$ [ $\sum_{k=1}^{n} f_{i k} g_{k j}$ ], then since $f_{i k} g_{k j}: M_{j} \rightarrow M_{i}$ thus the matrix product $\left[f_{i j}\right] \cdot\left[g_{i j}\right]$ is also an element of $H$-which shows that $H$ is indeed a ring.

If $\left[f_{i j}\right] \in H$ is any matrix, define a morphism of right- $R$-modules $f: M \rightarrow M$ with projections $\left(f_{i}\right)_{i=1}^{n}$, where for every $1 \leqslant i \leqslant n$ the map $f_{i}: M \rightarrow M_{i}$ is the unique morphism such that the diagram

commutes for all $1 \leqslant j \leqslant n$. We now simply define a map $\Phi: H \rightarrow \operatorname{End}_{\operatorname{Mod}_{R}}(M)$ by $\left[f_{i j}\right] \mapsto f$.

We now show that $\Phi$ is a ring morphism. Let $\left[g_{i j}\right] \in H$ be any other matrix and let $g:=\Phi\left(\left[g_{i j}\right]\right): M \rightarrow M$. If $\left[h_{i j}\right]:=\left[f_{i j}\right] \cdot\left[g_{i j}\right]$ then from definition we have $h_{i j}=\sum_{k=1}^{n} f_{i k} g_{k j}$. Notice that the composition $f g: M \rightarrow M$ with projections

$$
(f g)_{i}=f_{i}\left(\sum_{k=1}^{n} g_{k}\right)=f_{i}\left(\sum_{k=1}^{n} \sum_{j=1}^{n} g_{k j}\right)=\sum_{j=1}^{n} \sum_{k=1}^{n} f_{i k} g_{k j}
$$

for all $1 \leqslant i \leqslant n$, therefore $(f g)_{i}=h_{i}$. This proves that $\phi$ is a ring morphism:

$$
\phi\left(\left[f_{i j}\right] \cdot\left[g_{i j}\right]\right)=h=f g=\phi\left(\left[f_{i j}\right]\right) \phi\left(\left[g_{i j}\right]\right)
$$

For the injectivity of $\phi$ it suffices to see that the image of an element of $H$ is uniquely defined by the universal property of the coproduct. For the surjectivity of $\phi$, let $f: M \rightarrow M$ be any endomorphism, then again from the universal property each of its projections $f_{i}$ are uniquely defined by a family of morphisms of $R$-modules $\left(f_{i j}: M_{j} \rightarrow M_{i}\right)_{j=1}^{n}$, therefore, the matrix $\left[f_{i j}\right] \in H$ has image $f$ under $\phi$. Therefore we conclude that $H \simeq \operatorname{End}_{\text {Mod }_{R}}(M)$.

Example 8.4.64. Let $R$ be a ring and consider $R$ as a right- $R$-module over itself. We'll show that there exists a natural isomorphism of rings

$$
M_{n}(R) \simeq \operatorname{End}_{M_{o d}}\left(\bigoplus_{j=1}^{n} R\right)
$$

First, we shall show that $R \simeq \operatorname{End}_{\text {Mod }_{R}}(R)$ as rings. For that end, define a map $\Psi: R \rightarrow \operatorname{End}_{M_{M o d_{R}}}(R)$ given by $r \mapsto{ }_{r} m$, where ${ }_{r} m: R \rightarrow R$ is the left-multiplication by $r$, that is ${ }_{r} m(a)=r a$. It is easy to notice that if $r, s \in R$, then ${ }_{r s} m={ }_{r} m_{s} m$, therefore $\Psi$ is a ring morphism. For the injectivity of $\Psi$, if $r, s \in R$ are elements with equal image, then in particular $r={ }_{r} m(1)={ }_{s} m(1)=s$. For surjectivity, if $f: R \rightarrow R$ is an
endomorphism of right- $R$-modules, then given $r \in R$ we have $f(r)=f(1 \cdot r)=f(1) r$, therefore $f=f(1) m$-and $\Psi(f(1))=f$. We thus conclude that

$$
R \simeq \operatorname{End}_{\operatorname{Mod}_{R}}(R)
$$

As in Example 8.4.63, we can define a ring $H$ of $n \times n$ matrices whose entries are elements of the ring $\operatorname{End}_{\text {Mod }_{R}}(R)$. From our previous result, given a matrix $\left[f_{i j}\right] \in H$ there exists a unique matrix $\left[r_{i j}\right] \in M_{n}(R)$ such that $\Psi\left(r_{i j}\right)=f_{i j}$ for all $1 \leqslant i, j \leqslant n$. Therefore, $\Psi$ induces an isomorphism $H \simeq M_{n}(R)$. From Example 8.4.63 we know that $H \simeq \operatorname{End}_{M o d_{R}}\left(\bigoplus_{j=1}^{n} R\right)$, therefore we conclude that $M_{n}(R) \simeq \operatorname{End}_{M_{\text {Od }}^{R}}\left(\bigoplus_{j=1}^{n} R\right)$.

As a corollary of this result, together with Corollary 8.4.45, we find that if $S$ is a simple right- $R$-module, then there exists a natural isomorphism of rings

$$
M_{n}(D) \simeq \operatorname{End}_{M_{o d} R}\left(\bigoplus_{j=1}^{n} S\right)
$$

where $D$ is a division ring-in particular, we have $D \simeq \operatorname{End}_{\operatorname{Mod}_{R}}(S)$.

## Pullbacks \& Pushouts of $R$-Modules

Proposition 8.4.65 (Pullback). Let $M, N$, and $Z$ be $R$-modules. Given $R$-module morphisms $\mu: M \rightarrow Z$ and $\lambda: N \rightarrow Z$, we define a triple $\left(M \times_{Z} N, \pi_{M}, \pi_{N}\right)-$ where

$$
M \times_{Z} N:=\{(m, n) \in M \times N: \mu(m)=\lambda(n)\}
$$

with the natural $R$-module structures inherited from $M \times N$, and $\pi_{M}: M \times_{Z} N \rightarrow M$ and $\pi_{N}: M \times_{Z} N \rightarrow N$ are canonical projections. We claim that the following commutative square is a pullback


Proof. From construction we have $\mu \pi_{M}=\lambda \pi_{N}$, thus the diagram commutes. Let $W$ be any other $R$-module, and consider any two $R$-module morphisms $m: W \rightarrow M$ and $n: W \rightarrow N$ such that $\mu m=\lambda n$. We define a map $\phi: W \rightarrow M \times_{Z} N$ to be given by $\phi(w)=(m(w), n(w))$ - since $\mu m(w)=\mu n(w)$ then indeed $\phi(w) \in M \times_{Z} N$. Moreover, since $m$ and $n$ are morphisms, it follows trivially that $\phi$ is an $R$-module morphism. Uniqueness of $\phi$ comes from the fact that both $\pi_{M}$ and $\pi_{N}$ are epimorphisms.

Proposition 8.4.66 (Pushout). Let $M, N$, and $P$ be three $R$-modules, and consider morphisms $n: M \rightarrow N$ and $p: M \rightarrow P$. Consider the following:

- Define two $R$-modules:

$$
L:=\{(n(m),-p(m)): m \in M\} \subseteq N \oplus P \quad \text { and } \quad Q:=\frac{N \oplus P}{L}
$$

- Define morphisms $\varepsilon_{N}: N \rightarrow Q$ and $\varepsilon_{P}: P \rightarrow Q$ given by

$$
\varepsilon_{N}:=\pi \iota_{N} \quad \text { and } \quad \varepsilon_{P}:=\pi \iota \iota_{P} .
$$

Where $\pi: N \oplus P \rightarrow Q$ is the canonical projection and $\iota_{N}$ and $\iota_{P}$ are the canonical inclusions of $N$ and $P$, respectively, into $N \oplus P$.

Then the triple $\left(Q, \varepsilon_{N}, \varepsilon_{P}\right)$ is the pushout of the pair $(n, p)$, that is


Proof. First we show that $\varepsilon_{N} n=\varepsilon_{P} p$. Notice that for any $m \in M$ we have

$$
\begin{aligned}
\varepsilon_{P} p(m) & =\left(\pi \iota_{P}\right) p(m)=\pi(0, p(m))=(0, p(m))+L, \\
\varepsilon_{N} n(m) & =\left(\pi \iota_{N}\right) n(m)=\pi(n(m), 0)=(n(m), 0)+L,
\end{aligned}
$$

therefore $\varepsilon_{N} n(m)-\varepsilon_{P} p(m)=(n(m),-p(m))+L=L$, thus $\varepsilon_{N} n=\varepsilon_{P} p$.
Now, let $X$ be any $R$-module, and consider morphisms $p^{\prime}: P \rightarrow X$ and $n^{\prime}: N \rightarrow X$ such that $p^{\prime} p=n^{\prime} n$. Define a map $q: Q \rightarrow X$ given by $(x, y)+L \mapsto n^{\prime}(x)+p^{\prime}(y)$, where $(x, y) \in N \oplus P$. We now show that $q$ is well defined: suppose $(x, y)+L=\left(x^{\prime}, y^{\prime}\right)+L-$ that is, $\left(x-x^{\prime}, y-y^{\prime}\right) \in L$-then since $q\left(\left(x-x^{\prime}, y-y^{\prime}\right)+L\right)=q(L)=0$ and thus their images are the same. Also, from construction, $q$ is a uniquely determined morphism of $R$-modules and satisfies $p^{\prime}=q \varepsilon_{P}$ and $n^{\prime}=q \varepsilon_{N}$.

## Chapter 9

## Integral Domains

### 9.1 Torsion

Definition 9.1.1 (Torsion). Let $R$ be an integral domain, and $M$ be an $R$-module. We define the torsion of $M$ to be the submodule

$$
\text { tor } M:=\{m \in M: r m=0 \text { for some } r \in R \backslash 0\} .
$$

If tor $M=0$, we say that $M$ is of free-torsion. On the contrary, if tor $M=M$, then $M$ is said to be of torsion.

Definition 9.1.2 (Rank). Given an integral domain $R$ and an $R$-module $M$, we define the rank of $M$ to be the cardinality of the maximal $R$-linearly independent subset of $M$.

Example 9.1.3. If $R$ is an integral domain, and $M$ is a torsion $R$-module, then $\operatorname{rank}_{R} M=$ 0 . Indeed, for any singleton $\{m\} \subseteq M$ there exists a non-zero $r \in R$ such that $r m=0$, which shows that every singleton is $R$-linearly dependent.

### 9.2 Noetherian Rings \& Modules

## Chain Conditions

Recalling the definition of Noetherian rings (see Definition 8.3.35), we can extend this notion to the environment of modules:

Definition 9.2.1. Let $R$ be a commutative ring. We say that an $R$-module $M$ is Noetherian if every submodule of $M$ is finitely generated.

Proposition 9.2.2. Let $R$ be a commutative ring, and $M$ be an $R$-module. The following are equivalent propositions:
(a) $M$ is a Noetherian module.
(b) Every ascending chain of submodules of $M$ stabilizes. In other words, if $\left(N_{j}\right)_{j \in \mathbf{N}}$ is a collection of submodules of $M$ such that $N_{j} \subseteq N_{j+1}$, then there exists an index $j_{0} \in \mathbf{N}$ such that $N_{j}=N_{j+1}$ for all $j \geqslant j_{0}$.
(c) Every non-empty collection of submodules of $M$ has a maximal element with respect to inclusion.

Proof. - (a) $\Rightarrow$ (b). Let $M$ be Noetherian and define the module $N:=\bigcup_{j \in \mathbf{N}} N_{j}$, which is a submodule of $M$. Since submodules of Noetherian rings are finitely generated, it follows that $N$ is finitely generated. Let $N=\left\langle n_{1}, \ldots, n_{k}\right\rangle$ be its generating set. For all $1 \leqslant i \leqslant k$, there must exist $j_{i} \in \mathbf{N}$ such that $n_{i} \in N_{j}$ for all $j \geqslant j_{i}$. Taking the maximum $j_{0}:=\max \left(j_{1}, \ldots, j_{k}\right)$, one finds that $n_{i} \in N_{j}$ for each $1 \leqslant i \leqslant k$ and every $j \geqslant j_{0}$. Therefore $N \subseteq N_{j}$ for all $j \geqslant j_{0}$, which implies that $N_{j}=N$ for each of those indexes - therefore the chain stabilizes.

- (b) $\Rightarrow$ (c). We prove the contrapositive. Suppose there exists a non-empty collection $\mathcal{N}$ of submodules of $M$ admitting no maximal element. Let $N_{0} \in \mathcal{N}$ be any element. Inductively, for all $j \geqslant 1$, define $N_{j} \in \mathcal{N}$ such that $N_{j-1} \subsetneq N_{j}$, that is, $N_{j-1}$ is a proper subset of $N_{j}$ - this is possible since $N_{j-1}$ isn't maximal. The collection $\left(N_{j}\right)_{j \in \mathbf{N}}$ forms an ascending chain of submodules, but by construction does not stabilize.
- (c) $\Rightarrow$ (a). Let $N \subseteq M$ be any submodule. Since ( 0 ) $\subseteq N$ is a finitely generated submodule of $N$, one can define a non-empty collection $\mathcal{N}$ of finitely generated submodules of $N$. From (c) one has that $\mathcal{N}$ admits a maximum element, say $W:=\left\langle n_{1}, \ldots, n_{k}\right\rangle$. Let $n \in N$ be any element and consider the finitely generated submodule $\left\langle n_{1}, \ldots, n_{k}, n\right\rangle \in \mathcal{N}$. Since $W$ is maximal, we have $\left\langle n_{1}, \ldots, n_{k}, n\right\rangle \subseteq W$ - therefore $n \in W$ and $N \subseteq W$. Therefore $N=W$ is finitely generated, which proves that $M$ is Noetherian.

Corollary 9.2.3. Every principal ideal domain $R$ is a Noetherian module over itself, and thus every non-empty collection of ideals of $R$ admits a maximal element.

Lemma 9.2.4 (Quotient of Noetherian rings is Noetherian). Let $R$ be a Noetherian ring, and $\mathfrak{a} \subseteq R$ be an ideal. Then the quotient ring $R / \mathfrak{a}$ is Noetherian.

Proof. From Proposition 8.3 .36 we find that the canonical projection $R \rightarrow R / \mathfrak{a}$ implies that $R / a$ is Noetherian.

Theorem 9.2.5 (Generalized Hilbert's basis theorem). Let $R$ be a ring. Then $R$ is Noetherian if and only if the polynomial ring $R\left[x_{1}, \ldots, x_{n}\right]$.

## Prove generalized Hilbert's basis theorem

Corollary 9.2.6. Let $R$ be a Noetherian ring, and $\mathfrak{a} \subseteq R\left[x_{1}, \ldots, x_{n}\right]$ be an ideal of the polynomial ring. Then the quotient ring $R\left[x_{1}, \ldots, x_{n}\right] / \mathfrak{a}$ is Noetherian.

Proof. Since $R\left[x_{1}, \ldots, x_{n}\right]$ is Noetherian by Theorem 9.2.5, applying Lemma 9.2.4 we find that $R\left[x_{1}, \ldots, x_{n}\right] / \mathfrak{a}$ is Noetherian.

Corollary 9.2.7. Every finite-type algebra over a Noetherian ring is Noetherian.

## Existence of Maximal Ideals

Proposition 9.2.8. Let $R$ be a commutative ring. If $\mathfrak{a}$ is any proper ideal of $R$, then there exists a maximal ideal $\mathfrak{m}$ of $R$ containing $\mathfrak{a}$.

Proof. Consider the collection $I$ of proper ideals of $R$ containing a-which is ordered by inclusion. Using this ordering, let $\left(\mathfrak{a}_{j}\right)_{j \in J}$ be the chain of all proper ideals with $\mathfrak{a}_{j} \in I$ and $\mathfrak{a}_{j} \subseteq \mathfrak{a}_{j+1}$. Define the set $\mathfrak{m}:=\bigcup_{j \in J} \mathfrak{a}_{j}$, which is again an ideal of $R$. Since every $\mathfrak{a}_{j}$ contains $\mathfrak{a}$ and does not contain 1 , it follows that $\mathfrak{m}$ contains $\mathfrak{a}$ and also doesn't contain 1 -hence $\mathfrak{m}$ is a proper ideal. Therefore $\mathfrak{m}$ is a maximal ideal of $R$ containing $\mathfrak{a}$, which proves the statement.

### 9.3 Localization

Definition 9.3.1 (Multiplicative subset). Given a commutative ring $R$, a subset $S \subseteq R$ is said to be multiplicative if $1_{R} \in S$ and $S$ is closed under multiplication-that is, given $s, s^{\prime} \in S$, we have $s s^{\prime} \in S$.

Definition 9.3.2 (Localization). Let $R$ be a commutative ring, and $S \subseteq R$ be a multiplicative subset. A localization of $R$ over the set $S$ is a morphism of commutative rings $L_{S}: R \rightarrow R\left[S^{-1}\right]$ such that:
(a) For all $s \in S$, the image $L_{S}(s) \in R\left[S^{-1}\right]$ is a unit.
(b) For every morphism of rings $f: R \rightarrow K$ satisfying property (a), there exists a unique ring morphism $\phi: R\left[S^{-1}\right] \rightarrow K$ such that the diagram


Equivalently, consider the minimal equivalence relation $\sim_{\text {frac }}$ on the set $R \times S$ defined by $(r, s) \sim_{\text {frac }}\left(r^{\prime}, s^{\prime}\right)$ if and only if $r s^{\prime}-s r^{\prime}=0$. Then the localization of $R$ under the multiplicative subset $S$ is simply

$$
R\left[S^{-1}\right]=(R \times S) / \sim_{\text {frac }} .
$$

Example 9.3.3. Let $p \in \mathbf{Z}$ be a prime number, and consider the ring of fractions $\mathbf{Z}\left[p^{-1}\right]$. Then every proper submodule of the $\mathbf{Z}$-module given by the quotient $\mathbf{Z}\left[p^{-1}\right] / \mathbf{Z}$ is finite.

## How to solve this?

Definition 9.3.4. Let $R$ be an integral domain. We define the field of fractions of $R$ to be the localization

$$
\operatorname{Frac}(R):=R\left[(R \backslash\{0\})^{-1}\right] .
$$

It is immediate that $\operatorname{Frac}(R)$ is a field, since a field is simply an integral domain whose non-zero elements are units-and this exactly what we did with the above localization.

### 9.4 Principal Ideal Domains

## Modules Over PID's

Theorem 9.4.1. Let $R$ be a principal ideal domain, and $M$ be a free $R$-module with $\operatorname{rank}_{R} M=m$ finite, and $N \subseteq M$ a submodule. Then:
(a) The submodule $N$ is free with $\operatorname{rank}_{R} N=n$ satisfying $n \leqslant m$.
(b) There exists a basis $\left(y_{1}, \ldots, y_{m}\right)$ of $M$, and a collection $\left(a_{1}, \ldots, a_{m}\right)$ of $R$ such that $\left(a_{1} y_{1}, \ldots, a_{n} y_{n}\right)$ is a basis for $N$ and $a_{1}\left|a_{2}\right| \cdots \mid a_{n}$.

Proof. Fix a basis $\left(x_{1}, \ldots, x_{m}\right)$ for $M$ and define a collection of $R$-linear morphisms $\left(\pi_{j}: M \rightarrow R\right)_{j=1}^{n}$ with $\pi_{j} x_{i}=\delta_{i j}$.
(a) Since $R$ is a domain, $M$ is torsion-free and so is $N$. Since $M$ has a finite rank, it must also be the case that $N$ has a finite rank $n$ satisfying $n \leqslant m$-which follows from Proposition 10.2.14. We proceed via induction on $n$ : for the base case $n=0$ we have $N=0$ and therefore $N$ is free. We assume for the induction hypothesis that the proposition is true for some $0<n-1<m$. Now we consider the case $0<n \leqslant m$ : since $n$ is non-zero, then $N$ is a non-zero module and therefore there exists a non-zero $x=\sum_{j=1}^{n} b_{j} x_{j} \neq 0$ element of $N$. If $1 \leqslant j_{0} \leqslant n$ is an index such that $b_{j_{0}} \neq 0$, then in particular $\left.\pi_{j_{0}}\right|_{N}$ is a non-zero map of the form $N \rightarrow R$. Since $R$ is a PID and $\pi_{j_{0}} N \subseteq R$ is a non-zero submodule, then there must exist $b_{0} \in R \backslash 0$ for which $\pi_{j_{0}} N=R b_{0}$. Since $R$ is a domain, then $\left\{b_{0}\right\}$ is a basis for the free $R$-module $\pi_{j_{0}} N$. Then the short exact sequence

$$
0 \longrightarrow \operatorname{ker}\left(\left.\pi_{j_{0}}\right|_{N}\right) \longleftrightarrow N \xrightarrow{\left.\pi_{j_{0}}\right|_{N}} R b_{0} \longrightarrow 0
$$

ends with a free module and thus splits, hence

$$
N \simeq \operatorname{ker}\left(\left.\pi_{j_{0}}\right|_{N}\right) \oplus R b_{0}
$$

Since rank is additive over direct sums, then $n=\operatorname{rank}_{R}\left(\operatorname{ker}\left(\left.\pi_{j_{0}}\right|_{N}\right)\right)+1$ implies that $\operatorname{ker}\left(\left.\pi_{j_{0}}\right|_{N}\right) \subseteq M$ is a submodule of rank $n-1$. From the inductive hypothesis it follows that $\operatorname{ker}\left(\left.\pi_{j_{0}}\right|_{N}\right)$ is a free module. Therefore $N$ is the direct sum of two free modules, hence $N$ is itself free.
(b) Notice that the case $M=0$ is trivial, thus we shall consider only $m \geqslant 1$ and do induction on $m$-moreover, we'll assume that $N$ is anon-zero submodule of $M$, since the zero case is trivially satisfied. For the base case $m=1$, there exists an $R$ module isomorphism $\phi: M \xrightarrow{\simeq} R$. Then from the fact that $R$ is a PID it follows that there exists $a \in R$ such that $\phi N=R a=a R$, and if $u:=\phi^{-1}(1)$ then $M=R u$-so that $\{u\}$ is a basis for $M$. Moreover,

$$
N=a(R u)=(a R) u=(R a) u=R(a u)
$$

therefore $\{a u\}$ forms a basis for $N$.

Now let $m>1$ and assume as a hypothesis for induction that the proposition is true for all modules of rank less than $m$. For each $\phi \in \operatorname{Mor}_{R-\operatorname{Mod}}(M, R)$, let $a_{\phi} \in R$ denote a ring element such that $\phi N=R a_{\phi}$, and define a collection of principal ideals

$$
\Sigma:=\left(R a_{\phi}\right)_{\phi \in \operatorname{Mor}(M, R)} .
$$

From our previous considerations in item (a), we know that there must exist a non-zero $\pi_{j_{0}} \in \operatorname{Mor}(M, R)$-thus $\Sigma$ is non-empty and its elements are not all zero. Since $R$ is Noetherian, it follows that $\Sigma$ admits a maximal element $R a_{\phi_{1}} \in \Sigma$. Define $a_{1}:=a_{\phi_{1}}$ and let $y \in N$ be such that $\phi_{1} y=a_{1}$, which exists because $\phi_{1} N=R a_{1}$. We'll show that $a_{1} \mid \varphi y$ for any $\varphi \in \operatorname{Mor}(M, R)$. Let $\varphi: M \rightarrow R$ be any such map, and let $\mathfrak{p}:=\left\langle a_{1}, \varphi y\right\rangle$ be an ideal of $R$, so that there exists $d \in R$ such that $\mathfrak{p}=R d$ since $\mathfrak{p}$ is principal. Therefore there exists a pair $r_{1}, r_{2} \in R$ such that $d=r_{1} a_{1}+r_{2} \varphi y$. Define a morphism $\psi: M \rightarrow R$ given by $\psi:=r_{1} \phi_{1}+r_{2} \varphi$ so that $\psi y=d$ and hence $d \in \psi N$ from the fact that $y \in N$. Since $a_{1} \in R d$, then $R a_{1} \subseteq R d$ and from maximality we obtain the equality $R a_{1}=R d$, which proves that $a_{1} \mid d$-while $d \mid \varphi y$, thus $a_{1} \mid \varphi y$ as we wanted.
In particular, the last paragraph shows that $a_{1} \mid \pi_{j} y$ for each projection $1 \leqslant j \leqslant m$, thus we may find $b_{j} \in R$ such that $\pi_{j} y=a_{1} b_{j}$. We shall define

$$
y_{1}:=\sum_{j=1}^{m} b_{j} x_{j}
$$

Notice then that $y$ can be rewritten as

$$
\begin{equation*}
y=\sum_{j=1}^{m} \pi_{j}(y) x_{j}=\sum_{j=1}^{m}\left(a_{1} b_{j}\right) x_{j}=a_{1} \sum_{j=1}^{m} b_{j} x_{j}=a_{1} y_{1} \tag{9.1}
\end{equation*}
$$

Furthermore, one knows that $a_{1}=\psi y=\psi\left(a_{1} y_{1}\right)=a_{1} \psi y_{1}$, therefore since $R$ is a domain it follows that $\psi y_{1}=1$.

We'll now show that $M=R y_{1} \oplus \operatorname{ker} \psi$. Let $x \in M$ be any element and write $x=\psi(x) y_{1}+\left(x-\psi(x) y_{1}\right)$. Notice that

$$
\begin{equation*}
\psi\left(x-\psi(x) y_{1}\right)=\psi x-\psi(x) \psi\left(y_{1}\right)=0 \tag{9.2}
\end{equation*}
$$

so $x-\psi(x) y_{1} \in \operatorname{ker} \psi$. Therefore $M=R y_{1}+\operatorname{ker} \psi$. Now if $r y_{1} \in \operatorname{ker} \psi$ then from the fact that $\psi y_{1}=1$ we have $r=0$, hence $r y_{1}=0$ and thus $R y_{1} \cap \operatorname{ker} \psi=0$. We thus conclude that $M=R y_{1} \oplus \operatorname{ker} \psi$.
Further, we'll prove that $N=R\left(a_{1} y_{1}\right) \oplus(N \cap \operatorname{ker} \psi)$. Let $x \in N$ be any element and as before rewrite it as $x=\psi(x) y_{1}+\left(x-\psi(x) y_{1}\right)$. Using the fact that $\psi N=R a_{1}$, there exists $b \in R$ such that $\psi x=b a_{1}$-then one has

$$
x=\left(b a_{1}\right) y_{1}+\left(x-\left(b a_{1}\right) y_{1}\right)
$$

where $b a_{1} y_{1} \in R\left(a_{1} y_{1}\right)$ and $x-\left(b a_{1}\right) y_{1} \in N \cap \operatorname{ker} \psi$ from Eq. (9.2). Therefore $N=R\left(a_{1} y_{1}\right)+(N \cap \operatorname{ker} \psi)$, and since $R\left(a_{1} y_{1}\right) \subseteq R y_{1}$ while $N \cap \operatorname{ker} \psi \subseteq \operatorname{ker} \psi$ then $R\left(a_{1} y_{1}\right) \cap(N \cap \operatorname{ker} \psi)=0$. Thus

$$
N=R\left(a_{1} y_{1}\right) \oplus(N \cap \operatorname{ker} \psi)
$$

Since ker $\psi \subseteq M$ then it is free and the rank of the direct sum yields $\operatorname{rank}_{R} \operatorname{ker} \psi=$ $m-1$. Thus from the inductive hypothesis we conclude that there exists a basis $\left(y_{2}, \ldots, y_{m}\right)$ of $\operatorname{ker} \psi$ and a collection of ring elements $\left(a_{2}, \ldots, a_{n}\right)$ such that $\left(a_{2} y_{2}, \ldots, a_{n} y_{n}\right)$ forms a basis for $N \cap \operatorname{ker} \psi$, and $a_{2}\left|a_{3}\right| \cdots \mid a_{m}$. Since $M=$ $R y_{1} \oplus \operatorname{ker} \psi$ then $\left(y_{1}, y_{2}, \ldots, y_{m}\right)$ is a basis for $M$, and since $N=R\left(a_{1} y_{1}\right) \oplus(N \cap \operatorname{ker} \psi)$ we have that $\left(a_{1} y_{1}, a_{2} y_{2}, \ldots, a_{n} y_{n}\right)$ is a basis for $N$.
To conclude the proof, we must show that $a_{1} \mid a_{2}$. Since $\left(y_{1}, \ldots, y_{m}\right)$ is a basis of $M$, let $f: M \rightarrow R$ be an $R$-linear morphism such that $f y_{1}=f y_{2}=1$ while $f y_{j}=0$ for each $3 \leqslant j \leqslant m$. From Eq. (9.1) we have $a_{1}=a_{1} f y_{1}=f\left(a_{1} y_{1}\right)=f y$, therefore since $y \in N$ then $a_{1} \in f N$. This implies in $R a_{1} \subseteq f N$ but using maximality this yields the equality $f N=R a_{1}$. We also know that $a_{2}=f\left(a_{2} y_{2}\right)$ in a similar fashion, and since $a_{2} y_{2} \in N$ then $a_{2} \in f N$. This concludes the proof that $a_{1} \mid a_{2}$, which was the last step for the end of the proof at large.

## Decomposition Theorems

Theorem 9.4.2 (Fundamental theorem of invariant factors). Let $R$ be a principal ideal domain and $M$ a finitely generated $R$-module. Then the following holds:
(a) There exists a number $r \in \mathbf{N}$ and a collection of non-zero and non-invertible elements $\left(a_{1}, \ldots, a_{m}\right)$ of $R$ such that

$$
M \simeq R^{r} \oplus\left(\bigoplus_{j=1}^{m} R / R a_{j}\right)
$$

and $a_{1}\left|a_{2}\right| \cdots \mid a_{m}$. That is, $M$ is isomorphic to the direct sum of a finite collection of cyclic $R$-modules. The elements $a_{j}$ are said to be invariant factors of $M$.
(b) The module $M$ is torsion-free if and only if $M$ is free.
(c) There exists an isomorphism

$$
\operatorname{tor} M \simeq \bigoplus_{j=1}^{m} R / R a_{j}
$$

In particular, $M$ is a torsion module if and only if $r=0$, and in such a case we have Ann $M=R a_{m}{ }^{1}$.

[^13]Proof. Let $M=\left\langle x_{1}, \ldots, x_{n}\right\rangle$ and let $e_{j}:=\left(\delta_{i j}\right)_{i=1}^{n} \in R^{n}$ be a base element of $R^{n}$. Define a morphism $\phi: R^{n} \rightarrow M$ as the map that sends $e_{j} \mapsto x_{j}$ for each $j$-which is unique by the free module universal property. Since $\left\{x_{1}, \ldots, x_{n}\right\}$ generates $M$, then $\operatorname{im} \phi=M$ and $\phi$ is thus surjective. Via the first isomorphism theorem we obtain $R^{n} / \operatorname{ker} \phi \simeq M$. Since $\operatorname{ker} \phi$ is a submodule of $R^{n}$ we can use Theorem 9.4.1 to obtain that ker $\phi$ is free and has $\operatorname{rank}(\operatorname{ker} \phi):=m \leqslant n$. Furthermore, there exists a basis $\left(t_{1}, \ldots, t_{n}\right)$ of $R^{n}$ and a list of ring elements $\left(a_{1}, \ldots, a_{m}\right)$ such that $\left(a_{1} t_{1}, \ldots, a_{m} t_{m}\right)$ is a basis of $\operatorname{ker} \phi$ satisfying $a_{1}\left|a_{2}\right| \cdots \mid a_{m}$. From the decomposition of free modules we obtain

$$
M \simeq R^{n} / \operatorname{ker} \phi=\frac{R t_{1} \oplus \cdots \oplus R t_{n}}{R a_{1} t_{1} \oplus \cdots \oplus R a_{m} t_{m}}
$$

Define a morphism of $R$-modules $\psi: \bigoplus_{j=1}^{n} R t_{j} \rightarrow\left(\bigoplus_{j=1}^{m} R / R a_{j}\right) \oplus R^{n-m}$ by mapping

$$
\psi\left(r_{1} t_{1}, \ldots, r_{n} t_{n}\right):=\left(r_{1}+R a_{1}, \ldots, r_{m}+R a_{m}, r_{m+1}, \ldots, r_{n}\right) .
$$

From the definition, this map is certainly surjective. Moreover, we can see that

$$
\operatorname{ker} \psi=\left(\bigoplus_{j=1}^{m} R a_{j}\right) \oplus\{0\}^{\oplus(n-m)} \simeq \bigoplus_{j=1}^{m} R a_{j}
$$

therefore by the first isomorphism theorem we find

$$
\frac{R t_{1} \oplus \cdots \oplus R t_{n}}{R a_{1} t_{1} \oplus \cdots \oplus R a_{m} t_{m}} \simeq\left(\bigoplus_{j=1}^{m} R / R a_{j}\right) \oplus R^{n-m}
$$

Notice that if $a \in R$ is an invertible element, then $R a=R$ and hence $R / R a=0$. Therefore the proposition of item (a) is stablished since we can cut off of the direct sum every summand $R / R a_{j}$ where $a_{j}$ is invertible. Let $b \in R$ be any element, then given any $r+R b \in R / R b$ we find that

$$
b(r+R b)=b r+R b=r b+R b=R b
$$

therefore $R / R b$ is a torsion $R$-module. This shows that $M$ is torsion-free if and only if $M \simeq R^{r}$, which proves item (b). As we just noted, one has tor $M \simeq \bigoplus_{j=1}^{m} R / R a_{j}$, and $M$ is a torsion module we have Ann
Notice that this isn't finished: it is part of the lost work-this chapter was the most affected by the losses.

## Chinese Remainder Theorem

Theorem 9.4.3 (Chinese remainder). Let $R$ be a commutative ring, and $\mathfrak{a}_{1}, \ldots, \mathfrak{a}_{k}$ be ideals of $R$ such that $\mathfrak{a}_{i}+\mathfrak{a}_{j}=R$ for all $i \neq j$. Then:

- We have the equality $\mathfrak{a}_{1} \cap \cdots \cap \mathfrak{a}_{k}=\mathfrak{a}_{1} \cdot \ldots \cdot \mathfrak{a}_{k}$.
- The natural projection $R \rightarrow R / \mathfrak{a}_{1} \times \cdots \times R / \mathfrak{a}_{k}$ is surjective, and induces a natural isomorphism of rings

$$
\frac{R}{\mathfrak{a}_{1} \ldots \mathfrak{a}_{k}} \simeq\left(R / \mathfrak{a}_{1}\right) \times \cdots \times\left(R / \mathfrak{a}_{k}\right)
$$

Proof. In particular, we have $\mathfrak{a}_{j}+\mathfrak{a}_{k}=R$ for all $1 \leqslant j \leqslant k-1$. Therefore, for all such indices there exists $a_{j} \in \mathfrak{a}_{k}$ such that $1-a_{j} \in \mathfrak{a}_{j}$, hence

$$
\left(1-a_{1}\right) \ldots\left(1-a_{k-1}\right) \in \mathfrak{a}_{1} \cdot \ldots \cdot \mathfrak{a}_{k-1},
$$

and since $a_{j} \in \mathfrak{a}_{k}$, then $1-\prod_{j=1}^{k-1}\left(1-a_{j}\right) \in \mathfrak{a}_{k}$. This shows that

$$
\begin{equation*}
\left(\mathfrak{a}_{1} \cdot \ldots \cdot \mathfrak{a}_{k-1}\right)+\mathfrak{a}_{k}=R . \tag{9.3}
\end{equation*}
$$

Notice that $\mathfrak{a}_{1} \cdot \ldots \cdot \mathfrak{a}_{k} \subseteq \mathfrak{a}_{1} \cap \cdots \cap \mathfrak{a}_{k}$, thus it remains to prove the other side of the inclusion. From our last paragraph, we know that $\mathfrak{a}_{1} \cap \cdots \cap \mathfrak{a}_{k} \subseteq \mathfrak{a}_{1} \cdot \ldots \cdot \mathfrak{a}_{k}$ for all $k \geqslant 3$. Since $k=1$ is trivial, we just need to prove the case for $k=2$. Let $\mathfrak{b}, \mathfrak{c} \subseteq R$ be ideals such that $\mathfrak{b}+\mathfrak{c}=R$-then there exists $b_{0} \in \mathfrak{b}$ and $c_{0} \in \mathfrak{c}$ such that $b_{0}+c_{0}=1$. If $x \in \mathfrak{b} \cap \mathfrak{c}$, then $x=b_{0} x+c_{0} x$, implying in $x \in \mathfrak{b} \cdot \mathfrak{c}$. Thus $\mathfrak{b} \cap \mathfrak{c} \subseteq \mathfrak{b} \cdot \mathfrak{c}$.

We now prove the second assertion via induction on $k$. For the base case $k=1$, the statement follows trivially from the first isomorphism theorem. Assume as the hypothesis of induction that the statement is true $k-1>1$-that is, we have an isomorphism $R /\left(\mathfrak{a}_{1} \cdot \ldots \cdot \mathfrak{a}_{k-1}\right) \simeq\left(R / \mathfrak{a}_{1}\right) \times \cdots \times\left(R / \mathfrak{a}_{k-1}\right)$. Consider the natural map

$$
\begin{equation*}
\pi: R \longrightarrow \frac{R}{\mathfrak{a}_{1} \cdot \ldots \cdot \mathfrak{a}_{k-1}} \times\left(R / \mathfrak{a}_{k}\right) \tag{9.4}
\end{equation*}
$$

From Eq. (9.3) the statement is reduced for the case of two ideals $\mathfrak{b}:=\mathfrak{a}_{1} \cdot \ldots \cdot a_{k-1}$ and $\mathfrak{c}:=\mathfrak{a}_{k}$ such that $\mathfrak{b}+\mathfrak{c}=R$. Let $b, c \in R$ be any two elements-we shall show that there exists $r \in R$ such that $r-b \in \mathfrak{b}$ and $r-c \in \mathfrak{c}$. Since $\mathfrak{b}$ and $\mathfrak{c}$ are relatively prime, let as before $b_{0}+c_{0}=1$ and define $r:=b_{0} c+c_{0} b$. Then, one has

$$
\begin{array}{ll}
r=b_{0} c+\left(1-b_{0}\right) b=b+b_{0}(c-b) \equiv b & (\bmod \mathfrak{b}) \\
r=\left(1-c_{0}\right) c+c_{0} b=c+c_{0}(b-c) \equiv c & (\bmod \mathfrak{c})
\end{array}
$$

since $b_{0} \in \mathfrak{b}$ and $c_{0} \in \mathfrak{c}$. Therefore $r-c \in \mathfrak{b}$ and $r-b \in \mathfrak{c}$ a wanted. This shows that the natural morphism of rings $\pi$ (see Eq. (9.4)) is surjective. From our first considerations, we know that $\operatorname{ker} \pi=\mathfrak{b} \cap \mathfrak{c}=\mathfrak{b} \cdot \mathfrak{c}$, therefore the first isomorphism theorem establishes that $R /(\mathfrak{b} \cdot \mathfrak{c}) \simeq(R / \mathfrak{b}) \times(R / \mathfrak{c})$.

Corollary 9.4.4 (Chinese remainder for PIDs). Let $R$ be a principal ideal domain, and $a_{1}, \ldots, a_{k} \in R$ be elements such that $\operatorname{gcd}\left(a_{i}, a_{j}\right)=1$ for all pairs $i \neq j$. Then the natural map $r+\left(a_{1} \cdots a_{k}\right) \mapsto\left(r+\left(a_{1}\right), \ldots, r+\left(a_{k}\right)\right)$ establishes an isomorphism of rings

$$
R /\left(a_{1} \cdots a_{k}\right) \simeq \frac{R}{\left(a_{1}\right)} \times \cdots \times \frac{R}{\left(a_{k}\right)} .
$$

### 9.5 Unique Factorisation Domains

## Chapter 10

## Linear Algebra Over Rings

Remark 10.0.1 (Which side?). For the remainder of the chapter, if $M$ is said to be an $R$-module, then $M$ can be either a left or right $R$-module-and $R$ need not be commutativewe use this to generalize propositions for both left and right modules simultaneously, whenever possible. On the other hand, if we specify the side of the multiplicative structure of $M$, then it may well be the case that the proposition does only work for this strict case or that the distinction between right and left modules is important in some way.

Remark 10.0.2 (Algebras and commutative rings). We shall sometimes address a commutative ring by $k$-for instance, in the case of a $k$-algebra $A$.

### 10.1 Free Modules

## Construction

As always, if $S$ is a set and $M$ is an $R$-module, for some ring $R$, we define $M^{\oplus S}$ to be the collection of finitely supported set-functions $S \rightarrow M$. We define on $M^{\oplus S}$ an $R$-module structure as follows: for every $\alpha \in M^{\oplus S}$ and $r \in R$ we define

$$
(r \alpha)(s):=r(\alpha(s))
$$

If we consider the module of $R$ over itself, we can define a set-function $t: S \rightarrow R^{\oplus S}$ by mapping $s \mapsto \mathbf{s}$, where

$$
\mathbf{s}(x):= \begin{cases}1, & \text { if } x=s \\ 0, & \text { if } x \neq s\end{cases}
$$

For every set $S$, we define a corresponding module module $F_{R} S$ composed of of formal sums $\sum_{s \in S} a_{s} s$ such that $a_{s} \in R$ is non-zero only for finitely many $s \in S$.

Proposition 10.1.1. There exists a natural isomorphism of $R$-modules $F_{R} S \simeq R^{\oplus S}$.
Proof. Let $\Phi: F_{R} S \rightarrow R^{\oplus S}$ be a map given by $\sum_{s \in S} a_{S} S \mapsto \sum_{s \in S} a_{s}$ s. Clearly, such map is injective and preserves both multiplicative and additive structures, thus $\Phi$ is an $R$ module morphism. Moreover, if $\alpha \in R^{\oplus S}$ is any function, since $\alpha$ has finite support, the
formal sum $\sum_{s \in S} \alpha(s) s$ is a well defined element of $F_{R} S$. Also, mapping this element under $\Phi$ yields $\sum_{s \in S} \alpha(s) \mathbf{s}$ and, for every $x \in S$, we have $\sum_{s \in S} \alpha(s) \mathbf{s}(x)=\alpha(s) \mathbf{x}(x)=$ $\alpha(x)$. Therefore the map is surjective, hence an isomorphism.
Proof. Let $A$ and $B$ be two sets such that $F_{R} A=F_{R} B$. In particular, it follows that for all $a \in A$, there exists an element $a=\sum_{j=1}^{n} r_{j} b_{j} \in F_{R} B$, moreover, each $b_{i} \in B$ can be written as $\sum_{j=1}^{m} r_{j}\left(b_{i}\right) a_{j}$
Proposition 10.1.2 (Free module universal property). Let $R$ be a ring and $S$ be any set. Given any $R$-module $M$ and a set-function $f: S \rightarrow M$, there exists a unique $R$-module morphism $\phi: F_{R} S \rightarrow M$ such that the following diagram commutes in Set ${ }^{1}$ :

where $t: S \rightarrow F_{R} S$ is the mapping $s \mapsto s$.
Proof. Let $\phi: F_{R} S \rightarrow M$ be such that $\phi\left(\sum_{s \in S} a_{s} s\right) \mapsto \sum_{s \in S} a_{s} f(s)$ for any $\sum_{s \in S} a_{s} s \in$ $F_{R} S$ so that clearly $\phi \iota=f$ since $\phi \iota(s)=\phi(s)=f(s)$. Moreover, for any two $\sum_{s \in S} a_{s} s, \sum_{s \in S} b_{s} s \in F_{R} S$ we have

$$
\begin{aligned}
\phi\left(\sum_{s \in S} a_{s} s+\sum_{s \in S} b_{s} s\right) & =\phi\left(\sum_{s \in S}\left(a_{s}+b_{s}\right) s\right)=\sum_{s \in S}\left(a_{s}+b_{s}\right) f(s)=\sum_{s \in S} a_{s} f(s)+\sum_{s \in S} b_{s} f(s) \\
& =\phi\left(\sum_{s \in S} a_{s} s\right)+\phi\left(\sum_{s \in S} b_{s} s\right)
\end{aligned}
$$

Furthermore, $\phi$ certainly preserves the multiplicative structure by $R$ elements. Therefore $\phi$ is an $R$-module morphism. Since $f$ completely determines the image of $\phi$, it is the unique morphism such that $\phi \iota=f$.

Corollary 10.1.3. The mapping $t: S \hookrightarrow F_{R} S$ given by $s \mapsto s$ is injective.
Proof. For any pair $s, s^{\prime} \in S$ of distinct elements, consider the module $M:=F_{R}\left\{s, s^{\prime}\right\}$ and a set-function $f: S \rightarrow M$ with $f(s)=s$ and $f\left(s^{\prime}\right)=s^{\prime}$. Then, by the universal of free modules, there exists a unique morphism of $R$-modules $\phi: F_{R} S \rightarrow M$ such that $\phi \iota=f$. If $\iota(s)$ was equal to $\iota\left(s^{\prime}\right)$, then $f(s) f\left(s^{\prime}\right)$, which cannot be the case-thus $\iota(s) \neq \iota\left(s^{\prime}\right)$ for all $s, s^{\prime} \in S$ distinct, hence $\iota$ is injective.

Proposition 10.1.4. Let $A:=\{1, \ldots, n\}$ be a set of $n$ elements and define a map $\iota: A \rightarrow R\left[x_{1}, \ldots, x_{n}\right]$ by $j \mapsto x_{j}$. Then $R\left[x_{1}, \ldots, x_{n}\right]$ is a free commutative $R$-algebra on $A$.

Proof. We prove that $R\left[x_{1}, \ldots, x_{n}\right]$ satisfies the universal property for $A$. Let $\alpha: R \rightarrow M$ be any $R$-algebra and $f: A \rightarrow M$ be a set-function. From Proposition 8.2.9 we find a unique extension $\bar{\alpha}: R\left[x_{1}, \ldots, x_{n}\right] \rightarrow S$ such that $\left.\bar{\alpha}\right|_{R}=\alpha$ and $\bar{\alpha}\left(x_{j}\right):=f(j)$. Therefore, $\bar{\alpha}$ is the uniquely determined $R$-algebra morphism such that $\bar{\alpha} \iota=f$.

[^14]
## Free Modules from Subsets

Given any $R$-module $M$ and a subset $S \subseteq M$ of its elements, one can define a free module $R^{\oplus S}$ out of $S$ and, from the universal property of free modules, there exist a unique $R$-map $\phi: R^{\oplus S} \rightarrow M$ such that $\phi \iota=i$, where $\iota: S \hookrightarrow R^{\oplus S}$ and $i: S \hookrightarrow M$ is the canonical inclusion.

It is to be noted that the image $\phi\left(R^{\oplus S}\right) \subseteq M$ is a submodule of $M$ and its elements are of the form $\sum_{s \in S} a_{s} s$ for $a_{s} \in R$ non-zero only for finitely many $s \in S$. We'll denote the submodule generated by $S$ on $M$ by the notation $\langle S\rangle$.

Definition 10.1.5 (Finitely generated module). An $R$-module $M$ is said to be finitely generated exactly when there exists a finite subset $S \subseteq M$ such that $M=\langle S\rangle$. Equivalently, $M$ is finitely generated if there exists a surjective $R$-map $R^{\oplus n} \rightarrow M$ for a positive integer $n \in \mathbf{Z}_{>0}$.

Remark 10.1.6 (Submodules). The reader should be cautious when working with finitely generated modules, for instance, it is not true that a finitely generated module has finitely generated modules.

For instance, if we consider the ring of polynomials $P:=\mathbf{Z}\left[x_{1}, x_{2}, \ldots\right]$ on infinitely many variables, then $P=\langle 1\rangle$ is finitely generated as a $P$-module. However, if we consider the ideal $\mathfrak{a}:=\left(x_{1}, x_{2}, \ldots\right) \subseteq P$, one cannot take a finite collection of polynomials of $P$ and generate all of $\mathfrak{a}$. Indeed, if $S \subseteq P$ is any finite collection, since polynomials have finitely many terms, there must exist, for all $p\left(x_{1}, x_{2}, \ldots\right) \in S$, a maximum index $j \in \mathbf{Z}_{>0}$ such that $x_{j}$ appears as a variable in $p$ with a non-zero coefficient. Taking the maximum index over all polynomials of $S$, we are still left with only a finite index, say $n \in \mathbf{Z}_{>0}$, such that $x_{n}$ is the highest-index variable appearing any of the polynomials of $S$-hence $x_{n+1}$ is not contemplated by any of the polynomials of $S$, therefore this set cannot generate $\mathfrak{a}$.

Definition 10.1.7 (Noetherian module). An $R$-module $M$ is said to be a Noetherian module if every submodule of $M$ is finitely generated as an $R$-module.

We can state the definition for Noetherian rings (see Definition 8.3.35) equivalently as follows: a ring $R$ is said to be a Noetherian if $R$ is a Noetherian $R$-module.

Lemma 10.1.8. Let $M$ be an $R$-module, and $N \subseteq M$ be a submodule. Then if both $N$ and $M / N$ are finitely generated, then $M$ is finitely generated.

Proof. Let $N=\langle A\rangle$ and $M / N=\langle B\rangle$ for finite sets $A \subseteq N$ and $B \subseteq M / N$. If $m \in M$ is any element, then the class $m+N \in M / N$ can be written as $m+N=\sum_{b \in B} r_{b} b+N$ for $r_{b} \in R$. Moreover, since $m-\sum_{b \in B} r_{b} b \in N$, we can write $m-\sum_{b \in B} r_{b} b=\sum_{a \in A} r_{a} a$ for $r_{a} \in R$. Therefore

$$
m=\sum_{a \in A} r_{a} a+\sum_{b \in B} r_{b} b
$$

This shows that $A \cup B$ generates $M$, thus $M$ is finitely generated.
Proposition 10.1.9. Let $M$ be an $R$-module, and $N \subseteq M$ be a submodule. Then $M$ is Noetherian if and only if both $N$ and $M / N$ are Noetherian.

Proof. Analogous to the proof of Proposition 8.3.36, if $M$ is a Noetherian module, then the natural projection $\pi: M \rightarrow M / N$ shows that $M / N$ is Noetherian. Since $N$ is a submodule of $M$, then $N$ is finitely generated and every submodule $P \subseteq N$ is also a submodule of $M$, thus $P$ must be finitely generated.

For the converse, suppose both $M / N$ and $N$ are Noetherian. Fix any submodule $P \subseteq M$. Consider the submodule $P \cap N$ of both $P$ and $N$-from the hypothesis that $N$ is Noetherian, $P \cap N$ is finitely generated. By Proposition 8.4.53 we find that there exists a canonical isomorphism $P /(P \cap N) \simeq(N+P) / N$. Since $(N+P) / N$ is a submodule of $M / N$, by the hypothesis that $M / N$ is Noetherian it follows that $P /(P \cap N)$ is finitely generated. From Lemma 10.1 .8 we find that $P$ itself is finitely generated, making $M$ Noetherian.

Corollary 10.1.10. Given a Noetherian ring $R$, any finitely generated $R$-module $M$ is Noetherian.

Proof. Since $M$ is finitely generated, there exists a surjective $R$-module morphism $p: R^{\oplus n} \rightarrow M$, for some $n \in \mathbf{Z}_{>0}$ —therefore $M \simeq R^{\oplus n} / \operatorname{ker} p$. Now, if $R^{\oplus n}$ is Noetherian, by Proposition 10.1.9, then $\operatorname{ker} p$ and $R^{\oplus n} / \operatorname{ker} p$ are both Noetherian, and therefore $M$ is Noetherian.

We prove that $R^{\oplus n}$ is Noetherian by induction on $n$. For $n=1$ we have that $R^{\oplus 1} \simeq R$ is Noetherian. Assume that $R^{\oplus(n-1)}$ is Noetherian for some $n>1$. Notice that the inclusion $t: R^{\oplus(n-1)} \hookrightarrow R^{\oplus n}$ mapping $\left(r_{1}, \ldots, r_{n-1}\right) \mapsto\left(r_{1}, \ldots, r_{n-1}, 0\right)$ is an $R$-map and makes $R^{\oplus(n-1)}$ into a submodule of $R^{\oplus n}$. Considering the canonical projection of the $n$-th coordinate $\pi_{n}: R^{\oplus n} \rightarrow R$, we have ker $\pi_{n}=R^{\oplus(n-1)}$. By the first isomorphism theorem we obtain an isomorphism

$$
R^{\oplus n} / R^{\oplus(n-1)} \simeq R,
$$

therefore $R^{\oplus n} / R^{\oplus(n-1)}$ is Noetherian. Applying Proposition 10.1 .9 we find that $R^{\oplus n}$ is Noetherian-thus $M$ is Noetherian.

## Finite Generation of $R$-Algebras

Definition 10.1.11. Let $A$ be an $R$-algebra. We define the following two distinct concepts regarding the finiteness of the generation of $A$ :
(a) The $R$-algebra $A$ is said to be finite if there exists a surjective $R$-module morphism

$$
\phi: R^{\oplus n} \rightarrow A
$$

for some $n \in \mathbf{Z}_{>0}$, so that

$$
A \simeq R^{\oplus n} / \operatorname{ker} \phi,
$$

where $\operatorname{ker} \phi$ is a submodule of $R^{\oplus n}$. In other words, $A$ is finitely generated as a module over $R$. This nomenclature is unfortunately misleading: even though we say that $A$ is finite, $A$ may well be an infinite set.
(a) The $R$-algebra $A$ is said to be of finite type if there exists a surjective $R$-algebra morphism

$$
\alpha: R\left[x_{1}, \ldots, x_{n}\right] \rightarrow A
$$

for some $n \in \mathbf{Z}_{>0}$, so that

$$
A \simeq R\left[x_{1}, \ldots, x_{n}\right] / \operatorname{ker} \alpha
$$

where $\operatorname{ker} \alpha$ is an ideal of the ring $R\left[x_{1}, \ldots, x_{n}\right]$. In other words, $A$ is finitely generated as an algebra over $R$.

Remark 10.1.12 (Finite type but not finite). Let $A$ be an $R$-algebra. If $A$ is finite, then $A$ is also of finite type. On the other hand, if we consider the $R$-algebra $R[x]$, we find that $R[x]$ is clearly of finite type, but there exists no surjective $R$-map from $R^{\oplus n}$ to $R[x]$, therefore $R[x]$ isn't finite.

### 10.2 Linear Independence \& Bases

## Linear Independence

Definition 10.2.1 (Linear independence). Let $J$ be a set and $M$ be an $R$-module. An indexed set $j: J \rightarrow M$ is said to be linearly independent if the unique $R$-module morphism $\phi: F_{R} J \rightarrow M$, making the diagram

commute in Set, is injective. The indexed set $j$ is said to generate $M$ if $\phi$ is surjective. Finally, if $\phi$ is an isomorphism, then $j$ is a basis of $M$-in this case, since $F_{R} J \simeq R^{\oplus J}$, then $R^{\oplus J} \simeq M$.

Corollary 10.2.2. An $R$-module is free if and only if it admits a basis.
Proof. If $M$ is free, then there exists a set $S$ such that $M \simeq F_{R} S$, then $S$ is a basis for $M$. For the converse, if $M$ admits a basis $S$, then $F_{R} S \simeq M$.

As with vector spaces we'll intentionally identify an indexed set $j: J \rightarrow M$ simply by $J$ itself and say that $J$ is a subset of $M$. The good old abuse of notation.

Remark 10.2.3 (Singletons). Singletons do not need to be linearly independent. A simple example is $\{3\} \subseteq \mathbf{Z} / 9 \mathbf{Z}$.

The following lemma regarding maximality of linearly independent sets is equivalent to the Axiom of Choice.

Lemma 10.2.4 (Maximality). Let $M$ be an $R$-module and $S \subseteq M$ be a linearly independent set. There exists a maximal linearly independent subset of $M$ containing $S$.

Proof. Let $\mathcal{S}$ be the non-empty collection of linearly independent sets of $M$ containing $S$. Notice that a the union of a chain of elements of $\mathcal{S}$ is again a linearly independent set containing $S$-thus $\mathcal{S}$ is closed under arbitrary unions. In other words, every chain of elements has an upper bound in $\mathcal{S}$. By Zorn's lemma, it follows that the collection $\mathcal{S}$ has a maximal element.

Remark 10.2.5 (Maximality, generation, and bases). It should be noted right away that being a maximal linear independent set does not imply that the set generated the module. For instance, $\{2\} \subseteq \mathbf{Z}$ is a maximal linearly independent subset of $\mathbf{Z}$ containing $\{2\}$, but obviously it doesn't generate $\mathbf{Z}$. This, however is true for the case of vector spaces.

On the other hand, every base of a module is necessarily a maximal linearly independent set.

Definition 10.2.6 (Invariant basis number). A ring $R$ is said to satisfy the invariant basis number (IBN) property if $R^{m} \simeq R^{n}$ as $R$-modules if and only if $m=n$.

Corollary 10.2.7. A ring $R$ does not satisfy the invariant basis number property if and only if there exists distinct natural numbers $n, m \in \mathbf{N}$, and matrices $A \in M_{m \times n}(R)$ and $B \in M_{n \times m}(R)$ such that

$$
A B=\mathrm{id}_{m} \quad \text { and } \quad B A=\mathrm{id}_{n} .
$$

Corollary 10.2.8. Every field satisfies the invariant basis number property.
Proposition 10.2.9. Let $f: R \rightarrow S$ be a ring morphism, for any two rings $R$ and $S$. If $S$ satisfies the invariant basis number property, then so does $R$.

Proof. Let $A:=\left[a_{i j}\right] \in M_{m \times n}(R)$ and $B:=\left[b_{i j}\right] \in M_{n \times m}(R)$ be matrices and define both $f A:=\left[f\left(a_{i j}\right)\right] \in M_{m \times n}(S)$ and $f B:=\left[f\left(b_{i j}\right)\right] \in M_{n \times n}(S)$. Notice that since $f$ is a ring morphism we have $(f A)(f B)=f(A B)$. Since we cannot have $(f A)(f B)$ and $(f B)(f A)$ equal to their respective identities, it follows that $A B$ and $B A$ are also not equal to the identity matrices-therefore $R$ is satisfies the IBN property.

Theorem 10.2.10. All commutative rings satisfy the invariant basis number property.
Proof. Let $R$ be a commutative ring. By Proposition 9.2 .8 we find that $R$ contains a proper maximal ideal $\mathfrak{m}$-therefore the quotient ring $R / \mathfrak{m}$ is a field. Considering the canonical projection $R \rightarrow R / \mathfrak{m}$, we obtain, by Proposition 10.2 .9 , that $R$ is IBN since fields are IBN

Example 10.2.11. Let $V$ be an infinite dimensional $k$-vector space, for some field $k$. The ring $R:=\operatorname{End}_{\text {Vect }_{k}}(V)$ of endomorphisms of $V$ does not satisfy the IBN property.

First we prove that $\operatorname{End}_{\text {vect }_{k}}(V \oplus V) \simeq R^{4}$. Notice that any $k$-linear map $f: V \oplus V \rightarrow$ $V \oplus V$ can be decomposed into its projections $f_{1}, f_{2}: V \oplus V \rightarrow V$. Further, each projection $f_{j}$ can be again decomposed into unique $k$-linear morphisms $p_{j}, q_{j}: V \rightarrow V$ such that $f_{j}=p_{j}+q_{j}$-for $j \in\{1,2\}$. Therefore, the map $f$ is uniquely determined by the quadruple $\left(p_{1}, q_{1}, p_{2}, q_{2}\right)$, where $f=\left(p_{1}+q_{1}, p_{2}+q_{2}\right)$. Hence we may map bijectively $\operatorname{End}_{\text {Vect }_{k}}(V \oplus V) \rightarrow R^{4}$ via $f \mapsto\left(p_{1}, q_{1}, p_{2}, q_{2}\right)$.

Notice that, since $V$ is infinite dimensional, we have $\operatorname{dim}_{k} V=\operatorname{dim}_{k}(V \oplus V)$-thus $V \simeq V \oplus V$, and hence $\operatorname{End}_{\text {Vect }_{k}}(V \oplus V) \simeq \operatorname{End}_{\text {Vect }_{k}}(V)$ (see Proposition 8.4.21). Notice that from our earlier result, we just concluded that $R \simeq R^{4}$, which proves that $R$ does not satisfy the IBN property.

Example 10.2.12. Let $M:=\prod_{j \in \mathbf{N}} \mathbf{Z}$ be a Z-module and a ring $R:=\operatorname{End}_{Z-\mathrm{Mod}}(M)$. We'll show that $R$ does not satisfy the IBN property.

Define parallel Z-module morphisms $\phi, \psi: M \rightrightarrows M$ given by

$$
\begin{aligned}
& \phi\left(a_{1}, a_{2}, \ldots\right):=\left(a_{1}, a_{3}, a_{5}, \ldots\right) \\
& \psi\left(a_{1}, a_{2}, \ldots\right):=\left(a_{2}, a_{4}, a_{6}, \ldots\right) .
\end{aligned}
$$

Then, if $f \in R$ is any endomorphism, let $f_{j}: M \rightarrow \mathbf{Z}$ be its $j$-th projection. Define endomorphisms $g, h: M \rightrightarrows M$, whose projections are given by
$g_{2 j-1}:=f_{j}$ for odd $j \in \mathbf{Z}_{>0}$ and zero for the remaining projections,
$h_{2 j}:=f_{j}$ for even $j \in \mathbf{Z}_{>0}$ and zero for the remaining projections.
Then we obtain the equality

$$
\begin{aligned}
\phi g+\psi h & =\left(g_{1}, g_{3}, g_{5}, \ldots\right)+\left(h_{2}, h_{4}, h_{6}, \ldots\right) \\
& =\left(f_{1}, 0, f_{3}, 0, f_{5}, \ldots\right)+\left(0, f_{2}, 0, f_{4}, 0 f_{6}, \ldots\right) \\
& =f
\end{aligned}
$$

which is uniquely defined for $f$-hence $\{\phi, \psi\}$ is a basis for the right- $R$-module $R$ (over itself). From this we conclude that $R \simeq R^{2}$ by mapping $f \mapsto(g, h)$ and with an inverse $(g, h) \mapsto \phi g+\psi h=f$. This shows that $R$ does not satisfy the invariant basis number property.

## Free Modules Basis

Lemma 10.2.13. Let $R$ be an integral domain, and $M:=R^{\oplus A}$ be a free $R$-module. Considering the inclusion $M \hookrightarrow \operatorname{Frac}(R)^{\oplus A}$, a subset $S \subseteq M$ is linearly independent in $M$ if and only if it is linearly independent in $\operatorname{Frac}(R)^{\oplus A}$.

Proof. We do the proof of both statements via the contrapositive. Define the notation $V:=\operatorname{Frac}(R)^{\oplus A}$. Suppose $S$ is linearly independent in $V$, then there exists a collection of elements $\left(a_{s} / b_{s}\right)_{s \in S}$ for $a_{s} / b_{s} \in \operatorname{Frac}(R)$ non-zero only for finitely many $s \in S$-but not all zero-such that

$$
\begin{equation*}
\sum_{s \in S} \frac{a_{s}}{b_{s}} \cdot s=0 \tag{10.1}
\end{equation*}
$$

Consider the finite set $S^{\prime}:=\left\{s \in S: a_{s} \neq 0\right\}$. Since $b_{s} \neq 0$ for all $s \in S^{\prime}$, one can consider the non-zero finite product $b:=\prod_{s \in S^{\prime}} b_{s}$. Notice that, for all $s_{0} \in S$ we have

$$
b \cdot \frac{a_{s_{0}}}{b_{s_{0}}}=\left(\prod_{s \in S \backslash\left\{s_{0}\right\}} b_{s}\right) a_{s_{0}}
$$

which is simply an element of $R$ since the denominator of the fraction is 1 . We then define a collection $\left(c_{s}\right)_{s \in S}$ as $c_{s^{\prime}}:=\prod_{s \in S \backslash\left\{s^{\prime}\right\}} b_{s}$ for all $s^{\prime} \in S^{\prime}$, and $c_{s}:=1$ for all $s \in S \backslash S^{\prime}$. Notice that Eq. (10.1) is equivalent to

$$
\begin{equation*}
\sum_{s \in S}\left(c_{s} a_{s}\right) s=0 \tag{10.2}
\end{equation*}
$$

with coefficients $c_{s} a_{s} \in R$ for which finitely many are non-zero-but not all are zerotherefore Eq. (10.2) lies in $M$. From this we conclude that $S$ is linearly dependent on $M$.

For the converse, suppose that $S$ is linearly dependent on $M$, then there exists a collection $\left(r_{s}\right)_{s \in S}$-of elements $r_{s} \in R$ that are non-zero for only finitely many $s \in S$, and not all zero-such that

$$
\begin{equation*}
\sum_{s \in S} r_{s} s=0 \tag{10.3}
\end{equation*}
$$

From the inclusion $R \hookrightarrow \operatorname{Frac}(R)$, we see that $r_{s} \in \operatorname{Frac}(R)$ for all $s \in S$, and therefore Eq. (10.3) lies in both $M$ and $V$-which proves that $S$ is also linearly dependent in $V$.

Proposition 10.2.14. Let $R$ be an integral domain, and $M$ be a free $R$-module. If $B$ is a maximal linearly independent subset of $M$, and $S$ is a linearly independent subset of $M$, then

$$
|S| \leqslant|B| .
$$

Moreover, if $C$ is another maximal linearly independent subset of $M$, then

$$
|C|=|B| .
$$

Proof. By Lemma 10.2.13 it is equivalent to prove the proposition for a field $R=k$ and a $k$-vector space $M=V$.

We construct a map $\iota: S \rightarrow B$ inductively. Assume a well-ordering on the set $S$. We use transfinite induction-that is, assume $\iota$ is defined, injectively, for all $w<v$ for some $v \in S$, and define $B^{\prime}$ to consist of the elements of $B$ but for all $w<v$ we replace $\iota(w) \in B$ by $w$. Our inductive hypothesis will be that $B^{\prime}$ is still a maximal linearly independent set of $V$.

We now define $l(v)$. Since by hypothesis the set $B^{\prime}$ is maximal, then $B^{\prime} \cup\{v\}$ must be linearly dependent, so that there exists a collection $\left(a_{j}\right)_{j=0}^{n}$ of elements $a_{j} \in k$, not all zero, such that

$$
\begin{equation*}
a_{0} v+a_{1} b_{1}+\cdots+a_{n} b_{n}=0 \tag{10.4}
\end{equation*}
$$

for some finite subset $\left(b_{j}\right)_{j=1}^{n} \subseteq B^{\prime}$. Since $B^{\prime}$ is linearly independent, it must be the case that $a_{0} \neq 0$. Moreover, since $S$ is also linearly independent, it follows that not all $b_{j}$ can be members of $S$. With a possible change of indexing, we may thus assume that $a_{1} \neq 0$ and $b_{1} \in B^{\prime} \backslash S$ —this implies that $b_{1} \neq j(w)$ for all $w<v$, hence we may set $\iota(v):=b_{1}$ and not loose injectivity of $l$.

Consider now the set $B^{\prime \prime}$, whose elements are those of $B^{\prime}$ but with $b_{1}=\iota(v)$ replaced by $v$. Since $B^{\prime}$ is linearly independent, then $B^{\prime \prime} \backslash v$ is linearly independent therefore, by

Eq. (10.4) we find $v=-\sum_{j=1}^{n} \frac{a_{j}}{a_{0}} b_{j}$, hence no subset of $B^{\prime \prime}$-containing or not $v$-can be linearly dependent, since that would imply the linear dependence of $B^{\prime}$. We conclude that $B^{\prime \prime}$ is linearly independent and maximal, which finishes the transfinite induction and proves that the injective set-function $t: S \mapsto B$ can be constructed. This shows that $|S| \leqslant|B|$. For the equality, we can simply do the construction of injective maps $C \mapsto B$ and $B \mapsto C$ and conclude that $|C|=|B|$.

Corollary 10.2.15. Let $R$ be an integral domain, and $A$ and $B$ be sets. Then there exists an isomorphism of $R$-modules $F_{R} A \simeq F_{R} B$ if and only if there exists a bijection $A \simeq B$.

Proof. If $F_{R} A \simeq F_{R} B$, then $A$ and $B$ are maximal linearly independent sets of the "same" module, which by Proposition 10.2.14 implies in $|A|=|B|$.

If $|A|=|B|$, let $f: A \xrightarrow{\simeq} B$ be a bijective set function, and consider the canonical inclusions $i_{A}: A \hookrightarrow F_{R} A$ and $i_{B}: B \hookrightarrow F_{R} B$. Consider the indexed sets $j_{A}:=i_{B} f: A \rightarrow$ $F_{R} B$ and $j_{B}:=i_{A} f^{-1}: B \rightarrow F_{R} A$ both of which are injective, since are a composition of an injection with a bijection. The are unique induced $R$-module morphisms $\phi: F_{R} A \rightarrow$ $F_{R} B$ and $\psi: F_{R} B \rightarrow F_{R} A$ are, by construction, such that $\phi i_{A}=j_{A}$ and $\psi i_{B}=j_{B}$. Consider the following commutative diagram in Set:


Using the universal property on the first two squares we find that $\psi \phi$ is the unique $R$-module morphism induced by $i_{A}$-but since $\operatorname{id}_{F_{R} A}$ also satisfies the property, it follows that $\psi \phi=\operatorname{id}_{F_{R} A}$. Analogously, applying the universal property on the last two squares we obtain $\phi \psi=\operatorname{id}_{F_{R} B}$. Therefore $\phi$ and $\psi$ are inverses of each other and $F_{R} A \simeq F_{R} B$.

Lemma 10.2.16. Let $R$ be a commutative ring, and $F$ be a free $R$-module with a basis $B$. If $\mathfrak{m} \subseteq R$ is a maximal ideal, the collection $(b+\mathfrak{m} F)_{b \in B}$ forms a basis of the vector space $F /(\mathfrak{m} F)$ over the field $R / \mathfrak{m}^{2}$.

Proof. We shall prove that $B /(\mathfrak{m} F):=(b+\mathfrak{m} F)_{b \in B}$ forms a minimal generating set of $F /(\mathfrak{m} F)$ —so that by Lemma 5.6 .17 we obtain that $B /(\mathfrak{m} F)$ is a basis. If $x+\mathfrak{m} F \in F /(\mathfrak{m} F)$ is any element, consider a representative $x \in F$ and let $\left(x_{b}\right)_{b \in B}$ be a collection of elements $x_{b} \in R$-which are non-zero only for finitely many $b \in B$-such that $x=\sum_{b \in B} x_{b} b$. Taking the natural projection $F \rightarrow F /(\mathfrak{m} F)$ of sets, we find that $x+\mathfrak{m} F=\left(\sum_{b \in B} x_{b} b\right)+\mathfrak{m} F$.

[^15]By the $(R / \mathfrak{m})$-module structure of $F /(\mathfrak{m} F)$, we find that

$$
x+\mathfrak{m} F=\left(\sum_{b \in B} x_{b} b\right)+\mathfrak{m} F=\sum_{b \in B} x_{b} b+\mathfrak{m} F=\sum_{b \in B}\left(x_{b}+\mathfrak{m}\right)(b+\mathfrak{m} F) .
$$

From construction, the collection $\left(x_{b}+\mathfrak{m}\right)_{b \in B}$ is composed of finitely many non-zero elements $x_{b}+\mathfrak{m} \in R / \mathfrak{m}$. From this we conclude that $B /(\mathfrak{m} F)$ is indeed a generating set for $F /(\mathfrak{m} F)$.
How can I finish this proof? One can prove the linear independence or minimality of $B /(\mathrm{m} F)$, but what is the easiest way out?

Proposition 10.2.17. Let $R$ be a non-zero commutative ring. If $F$ is a free $R$-module and both $A$ and $B$ are basis of $F$, then $|A|=|B|$.

Proof. From Proposition 9.2 .8 we know that $R$ admits a maximal ideal $\mathfrak{m} \subseteq R$. We know from Lemma 10.2.16 that the collections $(a+\mathfrak{m} F)_{a \in A}$ and $(b+\mathfrak{m} F)_{b \in B}$ are both basis for the $(R / \mathfrak{m})$-vector space $F /(\mathfrak{m} F)$. Since basis of vector spaces have the same cardinality, it follows that $|A|=|B|$.

Definition 10.2.18 (Rank). Let $R$ be an integral domain. We define the rank of a free $R$-module $M$ to be the cardinality of a maximal linearly independent subset of $M$-this cardinal shall be denoted $\operatorname{rank}_{R} M$. If $R=k$ is a field, then $M$ is a $k$-vector space and thus $\operatorname{rank}_{k} M=\operatorname{dim}_{k} M$.

Proposition 10.2.19. Let $R$ be an integral domain and $M$ be an $R$-module. If $M$ is generated by a subset $S \subseteq M$, then $S$ contains a maximal linearly independent subset of $M$-hence $|S| \geqslant \operatorname{rank}_{R} M$.

Proof. By Lemma 10.2.13, we can treat the case for the field $k:=\operatorname{Frac}(R)$ and a vector space $V:=M$. Consider the power set $2^{S}$ of subsets of $S$, ordered by inclusionin particular, let $\mathcal{S} \subseteq 2^{S}$ contain all linearly independent subsets of $S$. By Zorn's lemma, there exists a maximal linearly independent set $B \in \mathcal{S}$ of $2^{S}$. Since maximal linearly independent sets form a basis (mind you, this is true only for vector spaces), then in particular $S$ is contained in the span of $B$-therefore $B$ is a maximal linearly independent set generating $V$, that is, $B$ is a basis for $V$.

Theorem 10.2.20 (Every module is the quotient of a free module). Given a ring $R$, every right- $R$-module $M$ (or left- $R$-module) is a quotient of a free right- $R$-module $F$ (or left- $R$-module). Moreover, $M$ is finitely generated if and only if one can choose a finitely generated free module $F$.

Proof. Define a free right- $R$-module $F:=R^{\oplus|M|}$ and let $\left(x_{m}\right)_{m \in M}$ be a basis of $F$. By the universal property of free modules, the indexing of the basis of $F$ induces a unique morphism of right- $R$-modules $p: F \rightarrow M$ such that $p\left(x_{m}\right)=m$ for all $m \in M$-which by construction is surjective. From the first isomorphism theorem we find that

$$
M \simeq F / \operatorname{ker} g
$$

therefore $M$ is indeed the quotient of a free module.
For the last proposition, if $F$ can be chosen to be finitely generated, then it is immediate that $M$ is finitely generated. For the converse, suppose that $M$ is finitely generated and $M=\left\langle m_{1}, \ldots, m_{n}\right\rangle$. Define the free $R$-module $F:=R^{\oplus n}$ with a basis $\left\{x_{1}, \ldots, x_{n}\right\}$ and consider the $R$-map $p: F \rightarrow M$ defined by $p\left(x_{j}\right):=m_{j}$ for each $1 \leqslant j \leqslant$ $n$. Since $\left\langle p\left(x_{1}\right), \ldots, p\left(x_{n}\right)\right\rangle=\left\langle m_{1}, \ldots, m_{n}\right\rangle$, then $p$ is surjective, therefore the statement holds true.

Corollary 10.2.21. For any $R$-module $M$ there exists an exact sequence of $R$-modules

$$
G \longrightarrow F \longrightarrow M \longrightarrow 0
$$

for free $R$-modules $G$ and $F$.

## Direct Sums \& Products of Free $R$-Modules

Proposition 10.2.22. Let $\left(M_{j}\right)_{j \in J}$ be a family of free $R$-modules. Then the direct sum $\bigoplus_{j \in J} M_{j}$ is free.
Proof. Let $\left(S_{j}\right)_{j \in J}$ be set of basis for each respective $M_{j}$, and define $S:=\bigcup_{j \in J} S_{j}$. Let $N$ be any $R$-module and $f: S \rightarrow N$ a set-function. Since $M_{j}$ is free for all $j \in J$, there exists a unique morphism $\phi_{j}: M_{j} \rightarrow N$ such that the diagram

commutes in Set. From the universal property of coproduct, the collection $\left(\phi_{j}\right)_{j \in J}$ defines a unique morphism $\phi: \bigoplus_{j \in J} M_{j} \rightarrow N$ such that the diagram

commutes in $R$-Mod for all $j \in J$. Therefore in particular

commutes in Set—which implies that $\bigoplus_{j \in J} M_{j}$ is free.
Remark 10.2.23. If $M$ is a free $R$-module and $N$ is a free submodule of $M$, it does not follow that the quotient $M / N$ is free. For instance, take $M:=\mathbf{Z}$ and the free submodule $N:=\mathbf{2 Z}$, then $M / N$ is not free since $\{1\}$ is not linearly independent in $\mathbf{Z} / 2 \mathbf{Z}$.

## Interesting Examples

Example 10.2.24. Let $R$ be an integral domain, and $M$ be a free $R$-module. If $a \in R$ and $m \in M$ are such that $a m=0$, than either $a=0$ or $m=0$.

Let $S$ be a basis for $M$ and consider $R$ as a module over itself. Let $f: S \mapsto R$ be any injective set-function. Since $t: S \hookrightarrow M$ is injective, then the unique $R$-module morphism $\phi: M \rightarrow R$ —making the diagram

commute in Set-must be injective, therefore $\operatorname{ker} \phi=0$. Notice that if $a m=0$ then since $\phi(0)=0$ we have $\phi(a m)=a \phi(m)=0$ but since $R$ is an integral domain, then either $a=0$ or $\phi(m)=0$-that is, $m \in \operatorname{ker} \phi$, which implies in $m=0$.

Example 10.2.25. Let $M$ be a free Z-module, and $v \in M$ a non-zero element. Then there exists only finitely many $n \in \mathbf{Z}$ such that the equation $v=n x$ has a solution $x \in M$.

Let $S$ be a basis for $M$, and let $v=\sum_{s \in S} a_{S} S$-where $a_{s} \in \mathbf{Z}$ is non-zero for only finitely many $s \in S$, but not all zero. Let $n \in \mathbf{Z}$ be an integer such that there exists a solution $x=\sum_{s \in S} b_{s} s \in M$ for $v=n x$, then since $S$ is a basis it follows that

$$
\sum_{s \in S}\left(a_{s}-n b_{s}\right) s=0
$$

implies in $a_{s}=n b_{s}$ for all $s \in S$. Hence $n$ must be a divisor of all $a_{s}$-thus be necessarily have $n \leqslant \operatorname{gcd}\left(a_{s}\right)_{s \in S}$. Since the greatest common divisor is finite, then $n$ can only assume a finite number of values for there to be a solution of $v=n x$ in $M$.

Remark 10.2.26. The direct product of a family of free modules need not be free. To see this, we construct the following example.

Let $M:=\prod_{j \in \mathbf{N}} \mathbf{Z}$, we'll show that $M$ is not a free $\mathbf{Z}$-module. Define the submodule $N:=\bigoplus_{j \in \mathbf{N}} \mathbf{Z}$ of $M$. Suppose, for the sake of contradiction, that $M$ admits a basis $B$.

Notice that since each element of $N$ has only finitely many non-zero entries, it follows that for each $n \in \mathbf{N}$ the of elements with $n$ non-zero entries can be represented by a sequence of pairs $\left(a_{j}, i_{j}\right)_{j=1}^{n}$ with $a_{j} \in \mathbf{Z}$ and a corresponding index $i_{j} \in \mathbf{N}$. Therefore, the number of elements with $n$ non-zero entries is given by the finite product of countable sets $(\mathbf{Z} \times \mathbf{N})^{n}$-which is itself countable. Then the number of elements of $N$ is given by the countable union $\bigcup_{n \in \mathbf{N}}\left(\bigcup_{j \in \mathbf{N}} \mathbf{Z}^{n}\right)$, which is again countable.

Let $B_{0} \subseteq B$ be the collection of all basis elements that appear when expanding the elements of $N$ in terms of $B$-such collection is necessarily countable since $N$ itself is countable. The free module $N_{0}:=F_{R} B_{0} \subseteq M$ will then contain $N$. Since $B_{0}$ is countable and the elements of $N_{0}$ are finite linear combinations of the elements of $B_{0}$, then $N_{0}$ is countable.

Consider the module $\bar{M}:=M / N_{0}$, has a basis $B \backslash B_{0}$. From Example 10.2.25, for every non-zero element $\bar{x} \in \bar{M}$ there exists only finitely many $n \in \mathbf{N}$ such that the equation $\bar{x}=n m$ admits a solution $m \in \bar{M}$.

Consider the subset

$$
S:=\left\{\left(b_{j}\right)_{j \in \mathbf{N}} \in M: b_{j}=( \pm j!) a_{j}, \text { with } a_{j} \in \mathbf{Z}\right\}
$$

which has cardinality of the continuum, $2^{\aleph_{0}}$, hence $S$ is uncountable. Therefore there exists $s \in S$ such that $s \notin N_{0}$. Notice however that, from the construction of $S$, every $n \in \mathbf{N}$ is such that the equation $s=n m$ has a solution $m \in M$-but since $s \notin N_{0}$, this also implies that $s=n \bar{m}$ has a solution $\bar{m} \in M / N_{0}$ for all $n \in \mathbf{N}$. This contradicts our last paragraph since $M / N_{0}$ is supposedly free! From this we conclude that $M / N_{0}$ must not be free, and $M$ should not admit a basis. Thus $M$ is not a free Z-module.

### 10.3 Exact Sequences \& Decompositions of Modules

## Direct Summands

Definition 10.3.1 (Direct summand). Let $P$ be an $R$-module and $M$ be a submodule of $P$. We say that $M$ is a direct summand of $P$ if there exists a submodule $N$ of $P$ satisfying $M \cap N=\{0\}$, called a complement of $M$, for which

$$
P=M \oplus N
$$

Definition 10.3.2 (Retract). Let $M$ be an $R$-module. A submodule $N$ of $M$ is said to be a retract of $M$ if there exists an $R$-module morphism $\rho: M \rightarrow N$ such that $\rho(n)=n$ for all $n \in N$-such morphism is called a retraction.

Proposition 10.3.3 (Direct summands \& retractions). Given $R$-modules $P$ and $M$, the module $M$ is a direct summand of $P$ if and only if there exists a retraction $\rho: P \rightarrow M$, in this case $P \simeq M \oplus \operatorname{ker} \rho$.

Proof. Suppose the retract $\rho$ exists. Let $p \in P$ be any element and consider the element $\rho(p) \in M$. Notice that $\rho(p-\rho(p))=\rho(p)-\rho(\rho(p))=\rho(p)-\rho(p)=0$, thus $p-\rho(p) \in \operatorname{ker} \rho$. Since $p=\rho(p)+(p-\rho(p))$, then $P=M+\operatorname{ker} \rho$. Moreover, if $s \in S \cap \operatorname{ker} \phi$ it follows that $\rho(s)=s$ and $\rho(s)=0$, therefore $s=0$-thus $S \cap \operatorname{ker} \phi=\{0\}$, and $P=M \oplus \operatorname{ker} \rho$.

For the converse, suppose that $M$ is a direct summand of $P$ and let $N$ be its complement. Since any element $p \in P$ can be written uniquely as a sum $p=m+n$ for $m \in M$ and $n \in N$, the map $\rho: P \rightarrow M$ given by $m+n \mapsto m$ is well defined and unique-also being an $R$-module morphism. Notice also that $\left.\rho\right|_{M}=\mathrm{id}_{M}$.
$\square$
Corollary 10.3.4. If $P=M \oplus N$ and $M \subseteq A \subseteq P$, then

$$
A=M \oplus(A \cap N)
$$

Proof. Let $\rho: P \rightarrow M$ be a retraction, so that $\operatorname{ker} \rho=N$. Since $M \subseteq A$, then $\left.\rho\right|_{A}$ is a retraction between $A$ and $M$, and $\operatorname{ker}\left(\left.\rho\right|_{A}\right)=A \cap N$. This shows that $A \cap N$ is a complement of $M$ in $A$ and hence $A=M \oplus(A \cap N)$.

Corollary 10.3.5. Let $M$ be an $R$-module, and $p: M \rightarrow M$ be an idempotent endomorphism (that is, $p^{2}=p$ ). Then there exists a canonical isomorphisms

$$
M \simeq \operatorname{im} p \oplus \operatorname{ker} p
$$

Proof. Notice that the induced map $\bar{p}: M \rightarrow \operatorname{im} p$ given by $\bar{p}(m)=p(m)$ is a morphism of $R$-modules such that $\bar{p}(\ell)=p(\ell)=\ell$ for all $\ell \in \operatorname{im} p$-since $p^{2}=p$. This implies that $\bar{p}$ is a retraction and from Proposition 10.3.3 it follows that $M \simeq \operatorname{im} p \oplus \operatorname{ker} p$.

Proposition 10.3.6. Let $M$ be an $R$-module and $\left(M_{j}\right)_{j=1}^{n}$ be a family of $R$-modules. Then

$$
M \simeq M_{1} \oplus \cdots \oplus M_{n}
$$

if and only if there exists a collection of $R$-module morphisms $\left(\phi_{j}: M \rightarrow M\right)_{j=1}^{n}$ such that $\operatorname{im} \phi_{j} \simeq M_{j}$ for all $j$, we have $\phi_{i} \phi_{j}=0$ for each pair $i \neq j$, and $\sum_{j=1}^{n} \phi_{j}=\operatorname{id}_{M}$.

Proof. Suppose $\psi: M_{1} \oplus \cdots \oplus M_{n} \xrightarrow{\simeq} M$ is an isomorphism, then every element of $m \in M$ can be uniquely written as a sum $m=\sum_{j=1}^{n} a_{j} \psi\left(\left(\delta_{i j} m_{j}\right)_{i=1}^{n}\right)$ for $a_{j} \in R$ and $m_{j} \in M_{j}$. Define, for $1 \leqslant j \leqslant n$, a morphism $\phi_{k}: M \rightarrow M$ given by

$$
\sum_{j=1}^{n} a_{j} \psi\left(\left(\delta_{i j} m_{j}\right)_{i=1}^{n}\right) \stackrel{\phi_{k}}{\longmapsto} a_{k} \psi\left(\left(\delta_{i k} m_{k}\right)_{i=1}^{n}\right)
$$

Clearly one has $\phi_{i} \phi_{j}=0$ for indices $i \neq j$, and $\phi_{1}+\cdots+\phi_{n}=\mathrm{id}_{M}$. Also, the $\operatorname{map} \Phi_{j}: \operatorname{im} \phi_{j} \rightarrow M_{j}$ given by $a_{j} \psi\left(\left(\delta_{i j} m_{j}\right)_{i=1}^{n}\right) \mapsto a_{j} m_{j}$ establishes an isomorphism $\operatorname{im} \phi_{j} \simeq M_{j}$.

For the converse, suppose that the collection of morphisms $\left(\phi_{j}\right)_{j=1}^{n}$ exist and satisfy the required properties. Let $N$ be any $R$-module, together with a family of morphisms $\left(\psi_{j}: N \rightarrow \operatorname{im} \phi_{j}\right)$. There exists a uniquely defined morphism $\eta: N \rightarrow M$ given by $n \mapsto \sum_{j=1}^{n} \psi_{j}(n)$ such that the following diagram

for all $1 \leqslant j \leqslant n$. This shows that $M$ satisfies the universal property for the product of the family $\left(M_{j}\right)_{j=1}^{n}$, hence $M \simeq M_{1} \oplus \cdots \oplus M_{n}$.

## Exact Sequences

Just like in the category of groups or vector spaces, we define an exact sequence of modules as follows.

Definition 10.3.7 (Exact sequence). A sequence $\left(d_{n}: M_{n} \rightarrow M_{n-1}\right)_{n \in \mathbf{Z}}$ of $R$-module morphisms is said to form an exact sequence

$$
\ldots \longrightarrow M_{n+1} \xrightarrow{d_{n+1}} M_{n} \xrightarrow{d_{n}} M_{n-1} \longrightarrow \ldots
$$

if for all $n \in \mathbf{Z}$ we have $\operatorname{im} d_{n+1}=\operatorname{ker} d_{n}$. In particular, a sequence is said to be exact in $M_{m}$ if im $d_{m+1}=\operatorname{ker} d_{m}$-thus an exact sequence is exact in each of its modules.

Proposition 10.3.8. Let $A, B$ and $C$ be $R$-modules. The following are properties concerning exact sequences:
(a) A sequence

$$
0 \longrightarrow A \xrightarrow{\phi} B
$$

is exact if and only if $\phi$ is injective.
(b) A sequence

$$
B \xrightarrow{\psi} C \longrightarrow 0
$$

is exact if and only if $\psi$ is surjective.
(c) A sequence

$$
0 \longrightarrow A \xrightarrow{\kappa} B \longrightarrow 0
$$

is exact if and only if $\kappa$ is an isomorphism.
Proof. (a) If the sequence is exact then $\operatorname{ker} \phi=0$ and $\phi$ is injective. For the converse, if $\phi$ is injective, then ker $\phi=0$ and since the image of the morphism $0 \rightarrow A$ must be zero, the exactness condition is satisfied.
(b) If the sequence is exact then $\operatorname{im} \psi=\operatorname{ker}(B \rightarrow 0)=B$ and $\psi$ is surjective. If on the contrary we have $\psi$ surjective, then again $\operatorname{im} \psi=B$, and since the kernel of the morphism $B \rightarrow 0$ is the whole module $B$, it follows that the sequence is exact.
(c) From the previous two items we have that $\kappa$ is both injective and surjective, thus a bijection-and since bijective morphisms of $R$-modules are isomorphisms, the statement follows.

Definition 10.3.9. An exact sequence of $R$-modules of the form

$$
0 \longrightarrow A \longrightarrow B \longrightarrow C \longrightarrow 0
$$

is said to be a short exact sequence. Moreover, we shall call such a sequence an extension of $A$ by $C$-we may sometimes name $B$ as the "extension".

Proposition 10.3.10 (Isomorphism theorems). We now restate the isomorphism theorems of $R$-modules in terms of exact sequences:
(a) If $0 \rightarrow A \stackrel{f}{\mapsto} B \stackrel{g}{\rightarrow} C \rightarrow 0$ is a short exact sequence, then there exists natural isomorphisms

$$
A \simeq \operatorname{im} f \quad \text { and } \quad B / \operatorname{im} f \simeq C
$$

(b) Let $M$ be an $R$-module, and $S$ and $T$ be both submodules of $M$. Then the following commutative diagram has short exact rows and the third vertical morphism is an isomorphism:

(c) Let $M$ be an $R$-module, and consider submodules $S$ and $T$ of $M$ such that $S \subseteq T$. Then there exists a short exact sequence

$$
0 \longrightarrow S / T \longmapsto \xrightarrow{\iota} M / T \xrightarrow{\pi} M / S \longrightarrow 0
$$

Proof. (a) Since $f$ is an injection, the induced morphism $f: A \xrightarrow{\simeq} \operatorname{im} f$ is an isomorphism. Moreover, since $\operatorname{ker} g=\operatorname{im} f$ and $g$ is a surjection, by the first isomorphism theorem we have $B / \operatorname{im} f \simeq C$.
(b) By the second isomorphism theorem, the mapping $S /(S \cap T) \rightarrow(S+T) / T$ given by $s+S \cap T \mapsto s+T$ is an isomorphism.
(c) Simply define $\iota$ as the inclusion $S / T \hookrightarrow M / T$ —so that im $\iota=S / T$ —and $\pi$ as the natural projection $M / T \rightarrow M / S$ mapping $m+T \mapsto m+S$. Since $\pi$ is surjective and, as argued in Proposition 8.4.54, $\operatorname{ker} \pi=S / T$. Therefore $\operatorname{ker} \pi=\operatorname{im} \iota$-the sequence is exact.

Proposition 10.3.11. Let $0 \rightarrow M_{j} \stackrel{\alpha_{j}}{\mapsto} N_{j} \xrightarrow{\beta_{j}} L_{j} \rightarrow 0$ be a short exact sequence of $R$-modules for all $1 \leqslant j \leqslant n$. Then the induced sequence

$$
0 \longrightarrow \bigoplus_{j=1}^{n} M_{j} \longmapsto \bigoplus_{j=1}^{n} N_{j} \xrightarrow{\beta} \bigoplus_{j=1}^{n} L_{j} \longrightarrow 0
$$

Proof. - We first prove the injectivity of $\alpha$. Notice that $\left(m_{j}\right)_{j} \in \operatorname{ker} \alpha$ if and only if $\alpha_{j}\left(m_{j}\right)=0$ —which implies in $m_{j}=0$ since $\alpha_{j}$ is injective-therefore $\left(m_{j}\right)_{j}=0$. This implies in ker $\alpha=0$, hence $\alpha$ is injective.

- For the surjectivity of $\beta$, let $\left(\ell_{j}\right)_{j} \in \bigoplus_{j=1}^{n} L_{j}$ be any element. Since each $\beta_{j}$ is surjective, there exists $n_{j} \in N_{j}$ such that $\beta_{j}\left(n_{j}\right)=\ell_{j}$. Now, if we consider the element $\left(n_{j}\right)_{j} \in \bigoplus_{j=1}^{n} N_{j}$, we obtain $\beta\left(n_{j}\right)_{j}=\left(\beta_{j}\left(n_{j}\right)\right)_{j}=\left(\ell_{j}\right)_{j}$.
- For the exactness in $\bigoplus_{j=1}^{n} N_{j}$ we prove that $\operatorname{im} \alpha=\operatorname{ker} \beta$. Let $\left(m_{j}\right)_{j} \in \bigoplus_{j=1}^{n}$ be any element, then by definition $\alpha\left(m_{j}\right)_{j}=\left(\alpha_{j}\left(m_{j}\right)\right)_{j}:=\left(n_{j}\right)_{j}$. Since each $n_{j} \in \operatorname{im} \alpha_{j}$ and $\operatorname{im} \alpha_{j} \subseteq \operatorname{ker} \beta_{j}$, it follows that $\beta_{j}\left(n_{j}\right)=0$. Then $\beta\left(n_{j}\right)_{j}=\left(\beta_{j}\left(n_{j}\right)\right)_{j}=0$, which implies in $\operatorname{im} \alpha \subseteq \operatorname{ker} \beta$.
For the converse inclusion, let $\left(n_{j}\right)_{j} \in \operatorname{ker} \beta$ be any element- then in particular we must have $n_{j} \in \operatorname{ker} \beta_{j}$. Since $\operatorname{ker} \beta_{j} \subseteq \operatorname{im} \alpha_{j}$ for all $1 \leqslant j \leqslant n$, then there exists
$m_{j} \in M_{j}$ such that $\alpha_{j}\left(m_{j}\right)=n_{j}$. Therefore one obtains $\alpha\left(m_{j}\right)_{j}=\left(\alpha_{j}\left(m_{j}\right)\right)_{j}=\left(n_{j}\right)_{j}$, which shows that $\operatorname{ker} \beta \subseteq \operatorname{im} \alpha$. Thus $\operatorname{im} \alpha=\operatorname{ker} \beta$.

Lemma 10.3.12. Consider the diagram

composed of short exact sequences of $R$-modules. Then $\kappa \alpha: A \rightarrow K$ is an isomorphism if and only if $\beta \delta: D \rightarrow B$ is an isomorphism.

Proof. From symmetry of the sequences we simply prove the forward implication. That is, suppose $\kappa \alpha$ is an isomorphism. For injectivity, let $d \in \operatorname{ker} \beta \delta$ be any element then $\delta(d) \in \operatorname{ker} \beta$. By exactness, there exists $a \in A$ such that $\alpha(a)=\delta(d)$. However, since $\delta(d) \in \operatorname{ker} \kappa$, then $\kappa \alpha(a)=0$-and by the injectivity of $\kappa \alpha$ we find that $a=0$, implying in $\delta(d)=0$. Since $\delta$ is also injective, then $d=0$, therefore $\operatorname{ker} \beta \delta=0$, proving that $\beta \delta$ is injective.

For surjectivity, take any $b \in B$ and, since $\beta$ is surjective, let $x \in X$ be such that $\beta(x)=b$. Consider the element $\kappa(x) \in K$-since $\kappa \alpha$ is surjective, there exists $a \in A$ for which $\kappa \alpha(a)=\kappa(x)$, thus $x-\alpha(a) \in \operatorname{ker} \kappa$. From exactness there must exist $d \in D$ for which $\delta(d)=x-\alpha(a)$. With this in hands we obtain

$$
\beta \delta(d)=\beta(x-\alpha(a))=\beta(x)-\beta \alpha(a)=\beta(x)=b,
$$

proving that $\beta \delta$ is surjective.

## Split Exact Sequences

Definition 10.3.13 (Split sequence). A short exact sequence

$$
0 \longrightarrow A \xrightarrow{\iota} B \xrightarrow{p} C \longrightarrow 0
$$

is said to be split if $p$ is a split epimorphism-that is, there exists a morphism of $R$-modules (a section of $p$ ) $s: C \mapsto B$ such that $p s=\operatorname{id}_{C}$.

Proposition 10.3.14 (Split sequence extension). If the short exact sequence

$$
0 \longrightarrow A>\stackrel{\iota}{\longrightarrow} B \xrightarrow{p} C \longrightarrow 0
$$

is split, then there exists a natural isomorphism of $R$-modules

$$
B \simeq A \oplus C
$$

Proof. Since the sequence is split, let $s: C \rightarrow B$ be the section of $p$. We'll show that $B=\operatorname{im} \iota \oplus \operatorname{ims}$. Let $b \in B$ be any element, then

$$
\begin{aligned}
p(b-s p(b)) & =p(b)-p(s p(b)) \\
& =p(b)-(p s) p(b) \\
& =p(b)-\operatorname{id}_{C} p(b) \\
& =0
\end{aligned}
$$

therefore $b-s p(b) \in \operatorname{ker} p$. Since the sequence is exact, $\operatorname{im} \iota=\operatorname{ker} p$, there exists $a \in A$ such that $l(a)=b-s p(b)$-thus $b=l(a)+s p(b)$. From this we conclude that $B=\operatorname{im} \iota+\operatorname{ims}$. For this to be a direct summand, it remains to prove that the intersection of the images is empty. Suppose that $x \in B$ is common to both images, so that there exists $a \in A$ and $c \in C$ such that $\iota(a)=x=s(c)$. Post-composing with $p$ we get $p(x)=p \iota(a)=0$, thus $x \in \operatorname{ker} p$. Moreover, $p s(c)=0$ but since $p s=\operatorname{id}_{C}$, then $c=0$-thus $s(c)=0$ and $x=0$. We conclude that

$$
B=\operatorname{im} \iota \oplus \operatorname{ims} s
$$

but $\operatorname{im} \iota \simeq A$ and $\operatorname{im} s \simeq C$-since both are injective maps-then $B \simeq A \oplus C$. Notice that this could also be extracted as a consequence of Corollary 10.3.5.

Remark 10.3.15. It should be emphasized that the converse of Proposition 10.3.14 does not hold in general. For instance, consider cyclic free groups $A:=\langle a\rangle$ and $B:=\langle b\rangle$ of orders 2 and 4 , respectively. Endowing such groups with the structure of $\mathbf{Z}$-modules, we can define morphisms $t: A \rightarrow B$ mapping $a \mapsto 2 b$, and $p: B \rightarrow A$ with $b \mapsto a$, then the sequence

$$
0 \longrightarrow A \xrightarrow{\iota} B \xrightarrow{p} A \longrightarrow 0
$$

is exact but does not split since $\mathrm{id}_{\mathbf{Z} / 4 \mathbf{Z}} \nsim \mathrm{id}_{\mathbf{Z} / 2 \mathbf{Z}} \oplus \mathrm{id}_{\mathbf{Z} / 2 \mathbf{Z}}$.
Proposition 10.3.16 (Equivalent definition for split sequences). A short exact sequence $0 \rightarrow A \xrightarrow{\iota} B \xrightarrow{p} C \rightarrow 0$ is split if and only if $\iota$ is a split monomorphism-that is, there exists a morphism $r: B \rightarrow A$ such that $r \iota=\operatorname{id}_{A}$. Moreover, if $s: C \rightarrow B$ is the section of $p$, then

$$
s p+\iota r=\mathrm{id}_{B} .
$$

Proof. Suppose the sequence splits and let $s: C \mapsto B$ be a section of $p$, then by Proposition 10.3.14 we have $B \simeq \operatorname{im} \iota \oplus \operatorname{ims}$. Since $\iota$ and $s$ are both injective maps, for every $b \in B$ there exists a unique pair $a \in A$ and $c \in C$ such that $b=\iota(a)+s(c)$. We define
a map $r: B \rightarrow A$ given by $r(b)=r(\iota(a)+s(c)):=a$. Therefore, for all $a \in A$, we have $r \iota(a)=a$-thus $r$ is a retract of $\iota$.

Suppose, for the converse, that $\iota$ is a split monomorphism with a retract $r: B \rightarrow A$. Notice that

$$
(\iota r)^{2}=(\iota r)(\iota r)=\iota(r \iota) r=\iota \mathrm{id}_{B} r=\iota r,
$$

therefore $\iota r: B \rightarrow B$ is an idempotent endomorphism. From Corollary 10.3.5 we obtain that $B=\operatorname{ker}(\iota r) \oplus \operatorname{im}(\iota r)$. Since $r$ is surjective, then $\operatorname{im}(\iota r)=\operatorname{im} \iota$. Moreover, since $p$ is surjective, given any $c \in C$, there exists $b \in B$ such that $p(b)=c$. From the decomposition of $B$, there exists $k \in \operatorname{ker}(\iota r)$ and $a \in A$ such that $b=k+\iota(a)$, therefore

$$
c=p(b)=p(k+\iota(a))=p(k)+p \iota(a)=p(k)
$$

since $\operatorname{im} \iota=\operatorname{ker} p$. Thus we may define a map $s: C \rightarrow B$ as $s(c)=s(p(k)):=k$, so that $s p=\operatorname{id}_{B}$. We conclude that $p$ is a split epimorphism, which implies that the sequence is split.

For the second statement, notice that

$$
\begin{aligned}
(s p+\iota r)(b) & =(s p+\iota r)(\iota(a)+s(c)) \\
& =(s p) f(a)+(s p) s(c)+(\iota r) \iota(a)+(\iota r) s(c) \\
& =s(p f(a))+s(p s(c))+\iota(r \iota(a))+\iota(r s(c)) \\
& =\iota(a)+s(c),
\end{aligned}
$$

therefore $s p+\iota r=\mathrm{id}_{B}$.
Proposition 10.3.17. If $F: R-\mathrm{Mod} \rightarrow \mathrm{Ab}$ is an additive functor of either variance, then there exists a natural isomorphism of abelian groups

$$
F(M \oplus N) \simeq F M \oplus F N
$$

In particular, if $F$ is covariant then the mapping $x \mapsto\left(\left(F \pi_{M}\right) x,\left(F \pi_{N}\right) x\right)$ is an isomorphismwhere $\pi_{M}: M \oplus N \rightarrow M$ and $\pi_{N}: M \oplus N \rightarrow N$ are natural projections.

Proof. Consider the projections $\pi_{M}, \pi_{N}$ and inclusions $\iota_{M}$ and $\iota_{N}$. From definition of the direct sum:
(a) Since $\pi_{M} \iota_{M}=\operatorname{id}_{M}$ and $\pi_{N} \iota_{N}=\operatorname{id}_{N}$ then

$$
\begin{aligned}
& F \mathrm{id}_{M}=\operatorname{id}_{F M}=F\left(\pi_{M} \iota_{M}\right)=F \pi_{M} F \iota_{M}, \\
& F \mathrm{id}_{N}=\operatorname{id}_{F N}=F\left(\pi_{N} \iota_{N}\right)=F \pi_{N} F \iota_{N} .
\end{aligned}
$$

(b) Since $\pi_{M} \iota_{N}=0: N \rightarrow M$ and $\pi_{N} \iota_{M}=0: M \rightarrow N$ we can use Proposition 8.4.19 to obtain that

$$
\begin{aligned}
& 0=F 0=F\left(\pi_{M} \iota_{N}\right)=F \pi_{M} F \iota_{N} \in \operatorname{Mor}_{\mathrm{Ab}}(F N, F M), \\
& 0=F 0=F\left(\pi_{N} \iota_{M}\right)=F \pi_{N} F \iota_{M} \in \operatorname{Mor}_{\mathrm{Ab}}(F M, F N) .
\end{aligned}
$$

(c) Since $\iota_{M} \pi_{M}+\iota_{N} \pi_{N}=\operatorname{id}_{M \oplus N}$, using the additivity of $F$ we obtain (for the case where $F$ is covariant)

$$
\begin{aligned}
\operatorname{id}_{F(M \oplus N)} & =F \operatorname{id}_{M \oplus N}=F\left(\iota_{M} \pi_{M}+\iota_{N} \pi_{N}\right)=F\left(\iota_{M} \pi_{M}\right)+F\left(\iota_{N} \pi_{N}\right) \\
& =F \iota_{M} F \pi_{M}+F \iota_{N} F \pi_{N}
\end{aligned}
$$

Therefore the maps $F \pi_{M}: F(M \oplus N) \rightarrow F M$ and $F \pi_{N}: F(M \oplus N) \rightarrow F N$ act as projections, while $F \iota_{M}: F M \rightarrow F(M \oplus N)$ and $F \iota_{N}: F N \rightarrow F(M \oplus N)$ are inclusions. Such maps satisfy all the requirements from Proposition 10.3.16 and therefore there exists a natural isomorphism $F(M \oplus N) \simeq F M \oplus F N$.

Proposition 10.3.18. Let $0 \rightarrow A \stackrel{f}{\mapsto} B \stackrel{g}{\rightarrow} F \rightarrow 0$ be a short exact sequence of $R$-modules. If $F$ is a free $R$-module, then the sequence is split. That is, every short exact sequence ending with a free module is split.

Proof. Let $\left(e_{j}\right)_{j \in J}$ be a basis for $F$. Since $g$ is surjective, let $\left(b_{j}\right)_{j \in J}$ be a collection such that $g\left(b_{j}\right)=e_{j}$. By the free module universal property, define the unique morphism $\rho: F \rightarrow B$ mapping $\rho\left(e_{j}\right) \mapsto b_{j}$ for each $j \in J$. Then notice that $g \rho\left(e_{j}\right)=g\left(b_{j}\right)=e_{j}$, thus we can again use the the uniqueness of the morphism of the universal property of free modules to obtain that $g \rho=\operatorname{id}_{F}$. Therefore we conclude that the sequence is split. 印

Example 10.3.19. Every short exact sequence of vector spaces is split. Indeed, any vector space admits a basis, which implies that a vector space is free-hence by Proposition 10.3.18 we obtain the proposition.

Proposition 10.3.20. If $0 \rightarrow A \stackrel{f}{\mapsto} B \stackrel{g}{\rightarrow} C \rightarrow 0$ is a short exact sequence of $R$-modules, then the following are properties concerning finite generation:
(a) If both $A$ and $C$ are finitely generated modules, then $B$ is finitely generated.
(b) If $B$ is finitely generated, then $C$ is finitely generated.

Proof. (a) Let $A:=\left\langle a_{1}, \ldots, a_{n}\right\rangle$ and $C:=\left\langle c_{1}, \ldots, c_{m}\right\rangle$. Define $\left(b_{j}^{\prime}\right)_{j=1}^{m}$ be a collection of elements $b_{j} \in g^{-1}\left(c_{j}\right)$ —which is ensured to exist by the surjectivity of $g$-and $\left(b_{j}\right)_{j=1}^{n}$ to be the collection $b_{j}:=f\left(a_{j}\right)$. We'll show that $X:=\left(b_{1}, \ldots, b_{n}, b_{1}^{\prime}, \ldots, b_{m}^{\prime}\right)$ is a generating set for $B$. Let $b \in B$ be any element, then since $C$ is finitely generated, we can write

$$
g(b)=\sum_{j=1}^{m} c_{j} r_{j}^{\prime}=\sum_{j=1}^{m} g\left(b_{j}^{\prime}\right) r_{j}^{\prime}=\sum_{j=1}^{m} g\left(b_{j}^{\prime} r_{j}^{\prime}\right)=g\left(\sum_{j=1}^{m} b_{j}^{\prime} r_{j}^{\prime}\right)
$$

for some collection of elements $r_{j}^{\prime} \in R$. Hence $x:=b-\sum_{j=1}^{m} b_{j}^{\prime} r_{j}^{\prime} \in \operatorname{ker} g$. Notice however that, since $A$ is generated by $\left(a_{j}\right)_{j=1}^{n}$, then in particular $\operatorname{im} f=\operatorname{ker} g$ is generated by $\left(b_{j}\right)_{j=1}^{n}$. Therefore there exists a collection $\left(r_{j}\right)_{j=1}^{n}$ of elements $r_{j} \in R$ such that

$$
b=\sum_{j=1}^{m} b_{j}^{\prime} r_{j}^{\prime}+\sum_{j=1}^{n} b_{j} r_{j}
$$

This shows that the finite set $X$-whose cardinality is $m+n$-generates the $R$ module $B$, proving the proposition.
(b) For item (b), one may notice that since the sequence is exact, then in particular $g: B \rightarrow C$ is surjective. Moreover, since $B$ is finitely generated, we let $B=\left\langle b_{1}, \ldots, b_{k}\right\rangle$. Given any $c \in C$ one has $b \in B$ such that $g(b)=c$. By the generating property, we have $b=\sum_{j=1}^{k} b_{j} r_{j}$ for a collection $\left(r_{j}\right)_{j=1}^{k}$ of ring elements $r_{j} \in R$. Therefore, by the morphism property of $g$ we obtain

$$
c=g(b)=g\left(\sum_{j=1}^{k} b_{j} r_{j}\right)=\sum_{j=1}^{k} g\left(b_{j}\right) r_{j} .
$$

This shows that $C$ is finitely generated by the collection $\left(g\left(b_{j}\right)\right)_{j=1}^{k}$.

## Morphisms of Exact Sequences

Definition 10.3.21 (Morphisms of short exact sequences). A morphism between short exact sequences of $R$-modules $0 \rightarrow A \rightarrow B \rightarrow C \rightarrow 0$ and $0 \rightarrow X \rightarrow Y \rightarrow Z \rightarrow 0$ is a triple ( $\alpha, \beta, \gamma$ ), of morphisms of $R$-modules making the following diagram commute


Definition 10.3.22 (Equivalent sequences). Two short exact sequences of $R$-modules $0 \rightarrow A \mapsto B \rightarrow C \rightarrow 0$ and $0 \rightarrow X \mapsto Y \rightarrow Z \rightarrow 0$ are said to be equivalent if $A=X$, $C=Z$ and there exists an isomorphism of $R$-modules Biso $Y$ such that the diagram

commutes in $R$-Mod.
Proposition 10.3.23. Let $A \xrightarrow{\alpha} B \xrightarrow{\beta} C \rightarrow 0$ and $X \xrightarrow{\chi} Y \stackrel{\gamma}{\rightarrow} Z \rightarrow 0$ be exact sequences. If there exists a surjective morphism $f: A \rightarrow X$ and an isomorphism $g: B \xrightarrow{\simeq} Y$ such that $g \alpha=\chi f$, then there exists a unique isomorphism $h: C \rightarrow Z$ making the diagram

commutative in $R$-Mod.

Proof. Let $c \in C$ be any element. Since $\beta$ is surjective, there exists $b \in B$ such that $\beta(b)=c$, thus we may define $h: C \rightarrow Z$ as the mapping $c \mapsto \gamma g(a)$. To see that $h$ is well defined, consider $b^{\prime} \in \beta^{-1}(c)$, then $\beta\left(b-b^{\prime}\right)=0$ and since $\operatorname{ker} \beta=\operatorname{im} \alpha$, there exists $a \in A$ such that $\alpha(a)=b-b^{\prime}$. Therefore

$$
\gamma g(b)-\gamma g\left(b^{\prime}\right)=\gamma g\left(b-b^{\prime}\right)=\gamma g \alpha(a)=\gamma \chi f(a)=0
$$

since $\operatorname{im} \chi=\operatorname{ker} \gamma$-thus $h$ is indeed well defined, is a $R$-module morphism, and makes the diagram commute. For the uniqueness of $h$, suppose $h^{\prime}: C \rightarrow Z$ is another morphism making the diagram commute-that is, $h^{\prime} \beta=\gamma g$. For any $c \in C$, let $b \in \beta^{-1}(c)$ and notice that

$$
h^{\prime}(c)=h^{\prime} \beta(b)=\gamma g(b)=h \beta(b)=h(b),
$$

therefore $h^{\prime}=h$.
We now show that $h$ is an isomorphism. Let $c \in \operatorname{ker} h$ be any element, and $b \in \beta^{-1}(c)$, then $0=h \beta(b)=\gamma g(b)$-therefore $g(b) \in \operatorname{ker} \gamma=\operatorname{im} \chi$, thus there exists $x \in X$ such that $\chi(x)=g(b)$. From the surjectivity of $f$, there exists $a \in A$ such that $f(a)=x$. Since the first square is commutative,

$$
g \alpha(a)=\chi f(a)=\chi(x)=g(b)
$$

but since $g$ is injective then $\alpha(a)=b$. Therefore we conclude that

$$
c=\beta(b)=\beta \alpha(a)=0
$$

since $\operatorname{ker} \beta=\operatorname{im} \alpha$. Hence ker $h=0$ and $h$ is injective.
For the surjectivity of $h$, let $z \in Z$ be any element. Since $\gamma$ is surjective, let $y \in Y$ be such that $\gamma(y)=z$, then from the surjectivity of $g$, we let $b \in B$ be an element such that $g(b)=y$, then from the commutativity of the second square we obtain

$$
h \beta(b)=\gamma g(b)=\gamma(y)=z
$$

thus $\beta(b) \in C$ has image $z$ under $h$, and $h$ is therefore surjective.
Proposition 10.3.24. Let $0 \rightarrow A \xrightarrow{\alpha} B \xrightarrow{\beta} C$ and $0 \rightarrow X \stackrel{\chi}{\rightarrow} Y \xrightarrow{\gamma} Z$ be exact sequences. If there exists morphisms $g: B \rightarrow Y$ and $h: C \rightarrow Z$ such that $\gamma g=h \beta$, then there exists a unique morphism $f: A \rightarrow X$ such that

commutes in $R$-Mod. Moreover, if both $g$ and $h$ are isomorphisms, then $f$ is an isomorphism.

Proof. Let $a \in A$ be any element and define $\alpha(a):=b$. Since the top row is exact we have $b \in \operatorname{ker} \beta$. From the commutativity of the second square, we find

$$
\gamma g(b)=h \beta(b)=h(0)=0
$$

therefore $g(b) \in \operatorname{ker} \gamma$. Since the bottom row is exact, there must exist $x \in X$ such that $\chi(x)=g(b)$. Define a map $f: A \rightarrow X$ by sending $a \mapsto x$ as described above. From the injectivity of $\alpha$ and $\chi$, one has that $\alpha^{-1}(b)=\{a\}$ and $\chi^{-1}(g(b))=\{x\}$, therefore $f$ is well defined. If $a, a^{\prime} \in A$, and $r \in R$ are any elements, defining $\alpha(a):=b$ and $\alpha\left(a^{\prime}\right):=b^{\prime}$ $\alpha\left(a r+a^{\prime}\right)=\alpha(a) r+\alpha\left(a^{\prime}\right)=b r+b^{\prime}$. Moreover, $g(b), g\left(b^{\prime}\right) \in \operatorname{ker} \gamma$ thus there exists $x, x^{\prime} \in X$ such that $\chi(x)=g(b)$ and $\chi\left(x^{\prime}\right)=g\left(b^{\prime}\right)$. Therefore

$$
f\left(a r+a^{\prime}\right)=x r+x^{\prime}=f(a) r+f\left(a^{\prime}\right)
$$

which shows that $f$ is a morphism of $R$-modules. For the commutativity of the first square we have, for any $a \in A$,

$$
g \alpha(a)=g(b)=\chi(x)=\chi f(a)
$$

For the second statement, suppose that both $g$ and $h$ are isomorphisms. Let $a \in$ ker $f$ be any element, then if $\alpha(a)=b$ we obtain by commutativity of the first square that

$$
g \alpha(a)=g(b)=\chi f(a)=\chi(0)=0
$$

but since $g$ is injective, then $b=0$. Therefore $\alpha(a)=0$, which implies in $a=0$, since $\alpha$ is injective. This shows that $\operatorname{ker} f=0$ and thus $f$ is injective. For surjectivity, let $x \in X$ be any element, and let $y:=\chi(x)$ so that $y \in \operatorname{ker} \gamma$ by the exactness of the bottom row. Since $g$ is surjective, let $b \in B$ such that $g(b)=y$. From the commutativity of the second square we have

$$
h \beta(b)=\gamma g(b)=\gamma(y)=0
$$

therefore $\beta(b) \in \operatorname{ker} h$. Since $h$ is injective, then $\beta(b)=0$ and thus $\beta \in \operatorname{im} \alpha$ from the exactness of the top row. Let $a \in A$ be such that $\alpha(a)=b$, then from the commutativity of the first square we get

$$
\begin{equation*}
\chi f(a)=g \alpha(a)=g(b)=y . \tag{দ}
\end{equation*}
$$

Since $\chi$ is injective and $\chi(x)=y$, then $f(a)=x$. Therefore $f$ is an isomorphism.
Proposition 10.3.25 (Five lemma). Consider the following commutative diagram in $R$-Mod, whose rows are exact:


The following properties hold:
(a) If $\phi_{2}$ and $\phi_{4}$ are surjective and $\phi_{5}$ is injective, then $\phi_{3}$ is surjective.
(b) If $\phi_{2}$ and $\phi_{4}$ are injective and $\phi_{1}$ is surjective, then $\phi_{3}$ is injective.
(c) If $\phi_{1}, \phi_{2}, \phi_{4}$ and $\phi_{5}$ are isomorphisms, then $\phi_{3}$ is an isomorphism.

Proof. Let $\alpha_{j}: A_{j} \rightarrow A_{j+1}$ and $\beta_{j}: B_{j} \rightarrow B_{j+1}$ for $1 \leqslant j \leqslant 4$ be the morphisms shown in the diagram. We prove each item:
(a) Let $b_{3} \in B_{3}$ be any element, and define $b_{4}:=\beta_{3}\left(b_{3}\right)$. Since $\phi_{4}$ is surjective, there exists $a_{4} \in A_{4}$ such that $\phi_{4}\left(a_{4}\right)=b_{4}$. From the exactness of the bottom row we have $b_{4} \in \operatorname{ker} \beta_{4}$, since $b_{4} \in \operatorname{im} \beta_{3}$. By the commutativity of the forth square we have

$$
\phi_{5} \alpha_{4}\left(a_{4}\right)=\beta_{4} \phi_{4}\left(a_{4}\right)=\beta_{4}\left(b_{4}\right)=0
$$

therefore $\alpha_{4}\left(a_{4}\right) \in \operatorname{ker} \phi_{5}$. Since $\phi_{5}$ is injective, then $a_{4} \in \operatorname{ker} \alpha_{4}$. By the exactness of the top row, there exists $a_{3} \in A_{3}$ such that $\alpha_{3}\left(a_{3}\right)=a_{4}$. From the commutativity of the third square we get

$$
\beta_{3} \phi_{3}\left(a_{3}\right)=\phi_{4} \alpha_{3}\left(a_{3}\right)=\phi_{4}\left(a_{4}\right)=b_{4}
$$

Since $\beta_{3}\left(b_{3}\right)=b_{4}=\beta_{3} \phi_{3}\left(a_{3}\right)$, then $\phi_{3}\left(a_{3}\right)-b_{3} \in \operatorname{ker} \beta_{3}$. Using again the exactness of the bottom row, there exists $b_{2} \in B_{2}$ such that $\beta_{2}\left(b_{2}\right)=\phi_{3}\left(a_{3}\right)-b_{3}$. Since $\phi_{2}$ is surjective, there exists $a_{2} \in A_{2}$ such that $\phi_{2}\left(a_{2}\right)=b_{2}$. From the commutativity of the second square one has

$$
\phi_{3} \alpha_{2}\left(a_{2}\right)=\beta_{2} \phi_{2}\left(a_{2}\right)=\beta_{2}\left(b_{2}\right)=\phi_{3}\left(a_{3}\right)-b_{3},
$$

which implies in $\phi_{3}\left(a_{3}-\alpha_{2}\left(a_{2}\right)\right)=b_{3}$. This proves the surjectivity of $\phi_{3}$.
(b) Let $a_{3} \in \operatorname{ker} \phi_{3}$ be any element. By the commutativity of the third square we obtain

$$
\phi_{4} \alpha_{3}\left(a_{3}\right)=\beta_{3} \phi_{3}\left(a_{3}\right)=\beta_{3}(0)
$$

thus $\alpha_{3}\left(a_{3}\right) \in \operatorname{ker} \phi_{4}$-but since $\phi_{4}$ is injective, then $a_{3} \in \operatorname{ker} \alpha_{3}$. From the exactness of the top row, there exists $a_{2} \in A_{2}$ such that $\alpha_{2}\left(a_{2}\right)=a_{3}$. By the commutativity of the second square we have

$$
\beta_{2} \phi_{2}\left(a_{2}\right)=\phi_{3} \alpha_{2}\left(a_{2}\right)=\phi_{3}\left(a_{3}\right)=0,
$$

hence $\phi_{2}\left(a_{2}\right) \in \operatorname{ker} \beta_{2}$. From the exactness of the bottom row there must exist $b_{1} \in B_{1}$ such that $\beta_{1}\left(b_{1}\right)=\phi_{2}\left(a_{2}\right)$. Since $\phi_{1}$ is surjective, let $a_{1} \in A_{1}$ be such that $\phi_{1}\left(a_{1}\right)=b_{1}$. Using the commutativity of the first square we obtain

$$
\phi_{2} \alpha_{1}\left(a_{1}\right)=\beta_{1} \phi_{1}\left(a_{1}\right)=\beta_{1}\left(b_{1}\right)=\phi_{2}\left(a_{2}\right),
$$

therefore $\phi_{2}\left(\alpha_{1}\left(a_{1}\right)-a_{2}\right)=0$. Since $\phi_{2}$ is injective, then $\alpha_{1}\left(a_{1}\right)=a_{2}$. From exactness of the top row, we have $a_{2} \in \operatorname{ker} \alpha_{2}$, but since $\alpha_{2}\left(a_{2}\right)=a_{3}$, then $a_{3}=0$. This shows that ker $\phi_{3}=0$, thus $\phi_{3}$ is injective.
(c) This last item is a direct consequence of the above items (a) and (b).

Proposition 10.3.26 ( $3 \times 3$ lemma). Consider the following commutative diagram in $R$-Mod, whose columns are exact:


Then, the following properties hold:
(a) If the bottom two rows are exact, then the top row is exact.
(b) If the top two rows are exact, then the bottom row is exact.

Proof. Let's diagram chase!
(a) First we show that $\operatorname{im} \alpha \subseteq \operatorname{ker} \beta$. Let $a \in A$ be any element and define $b:=\alpha(a)$, $x:=f_{1}(a)$ and $y:=\chi(x)$. From the commutativity of the top left square:

$$
g_{1} \alpha(a)=\chi f_{1}(a)=\chi(x)=y
$$

thus $g_{1}(b)=y$. Since $y \in \operatorname{im} \chi$, by the exactness of the middle row we obtain $y \in \operatorname{ker} \gamma$. From the commutativity of the top right square:

$$
h_{1} \beta(b)=\gamma g_{1}(b)=\gamma(y)=0
$$

therefore $\beta(b) \in \operatorname{ker} h_{1}$, and since $h_{1}$ is injective, we conclude that $\beta(b)=0$. Therefore $\operatorname{im} \alpha \subseteq \operatorname{ker} \beta$ as wanted.
For the final part, we show that $\operatorname{ker} \beta \subseteq \operatorname{im} \alpha$. Let $b \in \operatorname{ker} \beta$ be any element. From the commutativity of the top right square:

$$
\gamma g_{1}(b)=h_{1} \beta(b)=h_{1}(0)=0
$$

therefore $g_{1}(b) \in \operatorname{ker} \gamma$. From the exactness of the middle row there exists $x \in X$ such that $\chi(x)=g_{1}(b)$. From both the commutativity of the bottom left square and the exactness of the middle row:

$$
\gamma f_{2}(x)=g_{2} \chi(x)=g_{2} g_{1}(b)=0,
$$

hence $f_{2}(x) \in \operatorname{ker} \gamma$, but since $\gamma$ is injective we conclude that $x \in \operatorname{ker} f_{2}$. From the exactness of the left column there exists $a \in A$ such that $f_{1}(a)=x$. By the commutativity of the top left square:

$$
g_{1} \alpha(a)=\chi f_{1}(a)=\chi(x)=y
$$

Since $g_{1}$ is injective and $g_{1}(b)=y$, then $\alpha(a)=b$. Thus $\operatorname{ker} \beta \subseteq \operatorname{im} \alpha$.
(b) We show that $\operatorname{im} \lambda \subseteq \operatorname{ker} \sigma$. Let $\ell \in L$ be any element and define $s:=\gamma(\ell)$. Since $f_{2}$ is surjective, let $x \in X$ be such that $f_{2}(x)=\ell$. From the exactness of the middle row we have $y:=\chi(x) \in \operatorname{ker} \gamma$. By the commutativity of the bottom left square:

$$
g_{2} \chi(x)=\lambda f_{2}(x)=\lambda(\ell)=s,
$$

therefore $g_{2}(y)=s$. Now using the commutativity of the bottom right square:

$$
\sigma g_{2}(y)=h_{2} \gamma(y)=h_{2}(0)=0,
$$

then $\sigma(s)=0$, hence $\operatorname{im} \lambda \subseteq \operatorname{ker} \sigma$.
Finally, we show that $\operatorname{ker} \sigma \subseteq \operatorname{im} \lambda$. Let $s \in \operatorname{ker} \sigma$ be any element. Since $g_{2}$ is surjective, let $y \in Y$ be such that $g_{2}(y)=s$. From the commutativity of the bottom right square:

$$
h_{2} \gamma(y)=\sigma g_{2}(y)=\sigma(s)=0,
$$

therefore $\gamma(y) \in \operatorname{ker} h_{2}$. From the exactness of the right column, there exists $c \in C$ such that $h_{1}(c)=\gamma(y)$. Since $\beta$ is surjective, there exists $b \in B$ for which $\beta(b)=c$. Using the commutativity of the top right square:

$$
\gamma g_{1}(b)=h_{1} \beta(b)=h_{1}(c)=\gamma(y),
$$

hence $\gamma\left(y-g_{1}(b)\right)=0$. By the exactness of the middle row, we can find $x \in X$ such that $\chi(x)=y-g_{1}(b)$. Applying the commutativity of the bottom left square:

$$
\lambda f_{2}(x)=g_{2} \chi(x)=g_{2}\left(y-g_{1}(b)\right)=g_{2}(y)-g_{2} g_{1}(b)=g_{2}(y)=s,
$$

therefore $s \in \operatorname{im} \lambda$. Thus indeed $\operatorname{ker} \sigma \subseteq \operatorname{im} \lambda$.

Proposition 10.3.27. Consider the following commutative diagram in $R$-Mod:


Then the top row is exact if and only if the bottom row is exact.

Proof. $(\Rightarrow)$ Suppose the top row is exact. We prove that the bottom row is exact in two parts:

- We show that $\operatorname{im} \chi \subseteq \operatorname{ker} \gamma$. Let $x \in X$ be any element, and define $y:=\chi(x)$. Since $\phi_{1}$ is surjective, let $a \in A$ be such that $\phi_{1}(a)=x$. By the commutativity of the first square:

$$
\phi_{2} \alpha(a)=\chi \phi_{1}(a)=\chi(x)=y .
$$

From the exactness of the top row we have $b:=\alpha(a) \in \operatorname{ker} \beta$. Hence, by the commutativity of the second square:

$$
\gamma \phi_{2}(b)=\gamma(y)=\phi_{3} \beta(b)=\phi_{3}(0)=0,
$$

thus $y \in \operatorname{ker} \gamma$-which implies in im $\chi \subseteq \operatorname{ker} \gamma$.

- We now show that ker $\gamma \subseteq \operatorname{im} \chi$. Let $y \in \operatorname{ker} \gamma$ be any element. Since $\phi_{2}$ is surjective, let $b \in B$ be such that $\phi_{2}(b)=y$. From the commutativity of the second square:

$$
\phi_{3} \beta(b)=\gamma \phi_{2}(b)=\gamma(y)=0
$$

thus $\beta(b) \in \operatorname{ker} \phi_{3}$-but since $\phi_{3}$ is injective, then $b \in \operatorname{ker} \beta$. From the exactness of the top row we find $a \in A$ such that $\alpha(a)=b$. Using the commutativity of the first square:

$$
\chi \phi_{1}(a)=\phi_{2} \alpha(a)=\phi_{2}(b)=y,
$$

therefore $y \in \operatorname{im} \chi$ and $\operatorname{ker} \gamma \subseteq \operatorname{im} \chi$.
$(\Leftarrow)$ If on the contrary we assume that the bottom row is exact, since $\phi_{1}, \phi_{2}$ and $\phi_{3}$ are invertible, one can simply consider the following commutative diagram:


We can now apply the first part of the proof and conclude that the sequence of modules $0 \rightarrow A \rightarrow B \rightarrow C \rightarrow 0$ is exact.

Lemma 10.3.28 (Snake lemma). Consider the following commutative diagram in $R$-Mod:


There exists an exact sequence:


## Exact Functors

Definition 10.3.29. Let $F: R$-Mod $\rightarrow \mathrm{Ab}$ be a covariant functor, and $G: R-\mathrm{Mod}^{\mathrm{op}} \rightarrow \mathrm{Ab}$ be a contravariant functor. Consider a sequence of $R$-modules

$$
0 \longrightarrow A \xrightarrow{f} B \xrightarrow{g} C \longrightarrow 0
$$

We define the following notions:
(a) If $0 \rightarrow A \rightarrow B \rightarrow C$ is exact, the functor $F$ is said to be left exact if the induced sequence of abelian groups

$$
0 \longrightarrow F A \xrightarrow{F f} F B \xrightarrow{F g} F C
$$

is exact. Dually, $G$ is said to be right exact if

$$
G C \xrightarrow{G g} G B \xrightarrow{G f} G A \longrightarrow 0
$$

is an exact sequence.
(b) If $A \rightarrow B \rightarrow C \rightarrow 0$ is exact, the functor $F$ is said to be right exact if the induced sequence of abelian groups

$$
F A \xrightarrow{F f} F B \xrightarrow{F g} F C \longrightarrow 0
$$

is exact. Dually, $G$ is said to be left exact if

$$
0 \longrightarrow G C \xrightarrow{G g} G B \xrightarrow{G f} G A
$$

is an exact sequence.
(c) If $0 \rightarrow A \rightarrow B \rightarrow C \rightarrow 0$ is short exact, the functor $F$ is said to be exact if the induced sequence of abelian groups

$$
0 \longrightarrow F A \xrightarrow{F f} F B \xrightarrow{F g} F C \longrightarrow 0
$$

is exact. Analogously, $G$ is said to be exact if

$$
0 \longrightarrow G C \xrightarrow{G g} G B \xrightarrow{G f} G A \longrightarrow 0
$$

Proposition 10.3.30. Given any $R$-module $M$, the functors

$$
\operatorname{Mor}_{R-\operatorname{Mod}}(M,-): R-\operatorname{Mod} \longrightarrow \mathrm{Ab} \quad \text { and } \quad \operatorname{Mor}_{R-\operatorname{Mod}}(-, M): R-\operatorname{Mod}^{\mathrm{op}} \longrightarrow \mathrm{Ab}
$$

are both left-exact.

Proof. (a) We first show that $\operatorname{Mor}_{R-\operatorname{Mod}}(M,-)$ is exact. Let $0 \rightarrow A \stackrel{f}{\mapsto} B \xrightarrow{g} C$ be an exact sequence of $R$-modules, and consider the induced sequence of abelian groups

$$
0 \longrightarrow \operatorname{Mor}_{R-\operatorname{Mod}}(M, A) \stackrel{f_{*}}{\longrightarrow} \operatorname{Mor}_{R-\operatorname{Mod}}(M, B) \xrightarrow{g_{*}} \operatorname{Mor}_{R-\operatorname{Mod}}(M, C) .
$$

First we show that $f_{*}$ is injective. Let $\ell \in \operatorname{ker} f_{*}$ be any morphism, then by definition we have $f_{*}(\ell)=f \ell=0$. Since $f$ is injective, we have $\operatorname{ker} f=0$, which implies in $\operatorname{im} \ell=0$-hence $\ell=0$. This shows that $\operatorname{ker} f_{*}=0$, therefore $f_{*}$ is an injective map.
To show that the induced sequence is exact in $\operatorname{Mor}_{R-\operatorname{Mod}}(M, B)$, we prove that $\operatorname{im} f_{*}=\operatorname{ker} g_{*}$. Let $\ell \in \operatorname{im} f_{*}$ be any element and let $k \in \operatorname{Mor}_{R-\operatorname{Mod}}(M, A)$ be such that $f_{*}(k)=f k=\ell$. Since the original sequence is exact, we have

$$
g_{*}(\ell)=g \ell=g(f k)=0
$$

since $\operatorname{im} f=\operatorname{ker} g$. Therefore $\ell \in \operatorname{ker} g_{*}$ and $\operatorname{im} f_{*} \subseteq \operatorname{ker} g_{*}$.
For the converse of this inclusion, let $h \in \operatorname{ker} g_{*}$ be any map so that $g h(m)=0$ for all $m \in M$-therefore $h(m) \in \operatorname{ker} g$ and by exactness this means that there must exist $a \in A$ such that $f(a)=m$, which is unique since $f$ is an injective map. Define $p: M \rightarrow A$ by mapping $m \mapsto a$ if $h(m)=f(a)$. To check that $p$ is a morphism of modules, let $m, m^{\prime} \in M$ be elements and let $a, a^{\prime} \in A$ be such that $h(m)=f(a)$ and $h\left(m^{\prime}\right)=f\left(a^{\prime}\right)$. Then since

$$
h\left(m+m^{\prime}\right)=h(m)+h\left(m^{\prime}\right)=f(a)+f\left(a^{\prime}\right)=f\left(a+a^{\prime}\right)
$$

then $p\left(m+m^{\prime}\right)=a+a^{\prime}=p(m)+p\left(m^{\prime}\right)$. If $r \in R$ is any ring element, then one also has

$$
h(m r)=h(m) r=f(a) r=f(a r)
$$

therefore $p(m r)=a r=p(m) r$. Then we conclude that $p \in \operatorname{Mor}_{M o r} R(M, A)$ and therefore $f_{*}(p)=h$, showing that $h \in \operatorname{im} f_{*}$ and that $\operatorname{ker} g_{*} \subseteq \operatorname{im} f_{*}$. This finishes the proof that im $f_{*}=\operatorname{ker} g_{*}$.
(b) We now prove the left exactness of the contravariant functor $\operatorname{Mor}_{\operatorname{Mor} R}(-, M)$. Let $A \xrightarrow{i} B \xrightarrow{p} C \rightarrow 0$ be an exact sequence of $R$-modules, and consider the induced sequence of abelian groups

$$
0 \longrightarrow \operatorname{Mor}_{R-\operatorname{Mod}}(C, M) \stackrel{p^{*}}{\longrightarrow} \operatorname{Mor}_{R-\operatorname{Mod}}(B, M) \xrightarrow{i^{*}} \operatorname{Mor}_{R-\operatorname{Mod}}(A, M)
$$

We first show the injectivity of $p^{*}$. Let $h \in \operatorname{ker} p^{*}$ be any map, then for every $b \in B$ one has $h p(b)=0$, therefore $\operatorname{im} p \subseteq \operatorname{ker} h$. Since $p$ is surjective, $\operatorname{im} p=C$ and since ker $h \subseteq C$, then we conclude that $\operatorname{ker} h=\operatorname{im} p=C$ and therefore $h=0$.
Let's now show that $\operatorname{im} p^{*}=\operatorname{ker} i^{*}$. If $g \in \operatorname{Mor}_{R-\operatorname{Mod}}(C, M)$ is any map, then one has

$$
i^{*} p^{*}(g)=i^{*}(g p)=(g p) i=g(p i)=0
$$

from the fact that $\operatorname{im} i=\operatorname{ker} p$. Therefore $g \in \operatorname{ker} i^{*}$ and $\operatorname{im} p^{*} \subseteq \operatorname{ker} i^{*}$. For the converse, take $g \in \operatorname{ker} i^{*}$. Define a map $f: C \rightarrow Y$ by $c \mapsto g(b)$ if $c=p(b)$ —an element $b \in B$ with such property is ensured to exist by the surjectivity of $p$. To check that $f$ is well defined, let $b, b^{\prime} \in B$ be any two elements with $p(b)=p\left(b^{\prime}\right)$ then $p\left(b-b^{\prime}\right)=0$, which implies that there exists $a \in A$ such that $i(a)=b-b^{\prime}$. Moreover, since $i^{*}(g)=g i=0$, then

$$
g i(a)=g\left(b-b^{\prime}\right)=g(b)-g\left(b^{\prime}\right)=0,
$$

therefore $f(b)=f\left(b^{\prime}\right)$. To show that $f$ is a morphism of $R$-modules, let $r \in R$ and $c, c^{\prime} \in C$ be any elements with corresponding $b, b^{\prime} \in B$ for which $p(b)=c$ and $p\left(b^{\prime}\right)=c^{\prime}$. Then $p\left(b+b^{\prime}\right)=c+c^{\prime}$ and then

$$
f\left(c+c^{\prime}\right)=g\left(b+b^{\prime}\right)=g(b)+g\left(b^{\prime}\right)=f(c)+f\left(c^{\prime}\right)
$$

Moreover, we have $p(b r)=p(b) r=c r$, hence

$$
f(c r)=g(b r)=g(b) r=f(c) r .
$$

This shows that $f$ is indeed $R$-linear. Therefore we can finally note that $p^{*}(f)=$ $f p=g$ so that $g \in \operatorname{im} p^{*}$, hence $\operatorname{ker} i^{*} \subseteq \operatorname{im} p^{*}$.

Remark 10.3.31. The functors $\operatorname{Mor}_{R-\operatorname{Mod}}(M,-)$ and $\operatorname{Mor}_{R-M o d}(-, M)$ are not right-exact. In other words, if $0 \rightarrow A \xrightarrow{f} B$ is an exact sequence of $R$-modules, it is not always true that the sequence of abelian groups $\operatorname{Mor}_{R-\operatorname{Mod}}(A, M) \xrightarrow{f^{*}} \operatorname{Mor}_{R-\operatorname{Mod}}(B, M) \rightarrow 0$ is exact.

For instance, one can take the exact sequence $0 \rightarrow \mathbf{Z} \stackrel{f}{\mapsto} \mathbf{Z}$ where $f(x):=2 x$, and the module $M:=\mathbf{Z} / 2 \mathbf{Z}$, so that $\operatorname{Mor}_{\mathbf{Z}-\operatorname{Mod}}(\mathbf{Z}, M)=\{0, \pi\}$ where $\pi: \mathbf{Z} \rightarrow \mathbf{Z} / 2 \mathbf{Z}$ is the natural projection. Therefore the induced map $f^{*}: \operatorname{Mor}_{\mathbf{Z}-\operatorname{Mod}}(\mathbf{Z}, M) \rightarrow \operatorname{Mor}_{\mathbf{Z}-\operatorname{Mod}}(\mathbf{Z}, M)$ is not surjective, since $\pi \mapsto \pi f=0$.

Proposition 10.3.32 (Converse of Proposition 10.3.30). Consider $R$-modules $A, B$ and $C$, and morphisms of $R$-modules $f: A \rightarrow B$ and $g: B \rightarrow C$. We have the following two properties concerning the functors $\operatorname{Mor}_{R-\operatorname{Mod}}(M,-)$ and $\operatorname{Mor}_{R-\operatorname{Mod}}(-, M)$ :
(a) If the sequence of abelian groups

$$
0 \longrightarrow \operatorname{Mor}_{R-\operatorname{Mod}}(M, A) \stackrel{f_{*}}{\longrightarrow} \operatorname{Mor}_{R-\operatorname{Mod}}(M, B) \xrightarrow{g_{*}} \operatorname{Mor}_{R-\operatorname{Mod}}(M, C)
$$

is exact for every $M \in R$-Mod, then the corresponding sequence of $R$-modules

$$
0 \longrightarrow A \stackrel{f}{\longleftrightarrow} B \xrightarrow{g} C
$$

is also exact.
(b) If the sequence of abelian groups

$$
0 \longrightarrow \operatorname{Mor}_{R-\operatorname{Mod}}(C, M) \stackrel{g^{*}}{\longrightarrow} \operatorname{Mor}_{R-\operatorname{Mod}}(B, M) \xrightarrow{f^{*}} \operatorname{Mor}_{R-\operatorname{Mod}}(A, M) .
$$

is exact for every $M \in R$-Mod, then the corresponding sequence of $R$-modules

$$
A \xrightarrow{f} B \xrightarrow{g} C \longrightarrow 0
$$

is also exact.
Proof. (a) We first show the injectivity of $f$. Let $M:=\operatorname{ker} f$ and consider the natural inclusion $\iota \in \operatorname{Mor}_{R-\operatorname{Mod}}(M, A)$, then $f_{*}(\iota)=f \iota=0$, but since $f_{*}$ is injective then $\iota=0$-which implies in $\operatorname{ker} f=0$.
For the proof that $\operatorname{im} f=\operatorname{ker} g$, take $M:=\operatorname{im} f$ and consider the natural inclusion $i \in \operatorname{Mor}_{R-\operatorname{Mod}}(M, B)$. Since $f$ is injective, the induced map $\bar{f}: A \rightarrow \operatorname{im} f$ is a bijection, therefore one may define $\ell: \operatorname{im} f \rightarrow A$ by $\ell=(\bar{f})^{-1}$. Notice that $f_{*}(\ell)=f \ell=i$, therefore $i \in \operatorname{im} f_{*}$. Since $\operatorname{im} f_{*} \subseteq \operatorname{ker} g_{*}$, then $g_{*}(i)=g i=0$, which implies in $\operatorname{im} f \subseteq \operatorname{ker} g$.

For the converse inclusion, define $M:=\operatorname{ker} g$ and consider the canonical inclusion $j \in \operatorname{Mor}_{R-\operatorname{Mod}}(M, B)$, so that $g_{*}(j)=g j=0$. From the exactness of the sequence of abelian groups, there must exist $h \in \operatorname{Mor}_{R-\operatorname{Mod}}(M, A)$ such that $f_{*}(\ell)=f \ell=j$ therefore, for all $b \in \operatorname{ker} g$ there exists $\ell(b) \in A$ such that $f(\ell(b))=b$, thus $\operatorname{ker} g \subseteq$ $\operatorname{im} f$.
(b) First we prove that $g$ is surjective. Since the sequence is exact for all modules $M$, take $M:=C / \operatorname{im} g$ and let $\pi: C \rightarrow M$ be the natural projection morphism. Then we have $g^{*}(\pi)=\pi g=0$, but since $g^{*}$ is injective we find $\pi=0$-which can only be the case if $\operatorname{im} g=C$, therefore $g$ is surjective.

We shall now show that $\operatorname{im} f=\operatorname{ker} g$. Notice that since $\operatorname{im} g^{*} \subseteq \operatorname{ker} f^{*}$ then $f^{*} g^{*}=0$ but then $(g f)^{*}=0$ since $\operatorname{Mor}_{R-\operatorname{Mod}}(-, M)$ is contravariant. Therefore, if we take $M:=C$ and consider the element $\operatorname{id}_{C} \in \operatorname{Mor}_{R-\operatorname{Mod}}(C, M)$, we obtain

$$
0=(g f)^{*}\left(\mathrm{id}_{C}\right)=\operatorname{id}_{C}(g f)=g f,
$$

therefore $\operatorname{im} f \subseteq \operatorname{ker} g$.
For the other side of the inclusion, consider the module $M:=B / \operatorname{im} f$ and let $p: B \rightarrow M$, then denote the natural projection morphism—so that $f^{*}(p)=p f=0$. Since the sequence of abelian groups is exact, there must exist $h \in \operatorname{Mor}_{R-\operatorname{Mod}}(C, M)$ such that $g^{*}(h)=h g=p$. If, for the sake of contradiction, ker $g$ is not contained in $\operatorname{im} f$, there must exist $b \in B$ such that $g(b)=0$ but $b \notin \operatorname{im} f$. This however implies that $p(b) \neq 0$, while $h g(b)=p(b)$-and since $g(b)=0$ then $h g(b)=0$, which is a contradiction. This shows that $\operatorname{ker} g \subseteq \operatorname{im} f$.

Lemma 10.3.33. Consider the following sequence of $R$-modules: sequence of $R$ modules

$$
\begin{equation*}
0 \longrightarrow A \xrightarrow{f} B \xrightarrow{g} C \longrightarrow 0 \tag{10.5}
\end{equation*}
$$

We have the following:
(a) The sequence Eq. (10.5) is split short exact if and only if for every $R$-module $X$ the induced sequence of abelian groups

$$
0 \longrightarrow \operatorname{Mor}_{R-\operatorname{Mod}}(X, A) \stackrel{f_{*}}{\longrightarrow} \operatorname{Mor}_{R-\operatorname{Mod}}(X, B) \xrightarrow{g_{*}} \operatorname{Mor}_{R-\operatorname{Mod}}(X, C) \longrightarrow 0
$$

is exact.
(b) The sequence Eq. (10.5) is split short exact if and only if for every $R$-module $Y$ the induced sequence of abelian groups

$$
0 \longrightarrow \operatorname{Mor}_{R-\operatorname{Mod}}(C, Y) \stackrel{g^{*}}{\longrightarrow} \operatorname{Mor}_{R-\operatorname{Mod}}(B, Y) \xrightarrow{f^{*}} \operatorname{Mor}_{R-\operatorname{Mod}}(A, Y) \longrightarrow 0
$$

is exact.
Proof. We shall only prove item (a) since (b) is merely its dual. $(\Rightarrow)$ Since $\operatorname{Mor}_{R-\operatorname{Mod}}(X,-)$ is left exact, it suffices to show that $g_{*}$ is an epimorphism. Since $g$ is a retract, there exists $u: C \rightarrow B$ such that $g u=\operatorname{id}_{C}=g_{*} u$, therefore given any morphism $t \in \operatorname{Mor}_{R-M o d}(X, C)$ there exists a morphism $u t \in \operatorname{Mor}_{R-\operatorname{Mod}}(X, B)$ such that $g_{*}(u t)=g u t=t$-which shows that $g_{*}$ is surjective.
$(\Leftarrow)$ By Proposition 10.3 .32 we know that $0 \rightarrow A \rightarrow B \rightarrow C$ is exact. If we now choose $X:=C$ then there exists a map $v: C \rightarrow B$ such that $g_{*} v=g v=\mathrm{id}_{C}$ from the fact that $g_{*}$ is an epimorphism-thus surjective. This shows that $g$ admits a right inverse, thus $g$ is a split epimorphism.

Theorem 10.3.34. Let $A$ and $B$ be a pair of $k$-algebras, where $k$ is a commutative ring, and consider a pair of functors $F: A$-Mod $\rightarrow B$-Mod : $G$. If the pair $(F, G)$ is adjunct, then $F$ is right-exact and $G$ is left exact.

Proof. By definition, for any pair of modules $M \in A$-Mod and $X \in B$-Mod there exists a natural isomorphism

$$
\theta_{M X}: \operatorname{Mor}_{B-\operatorname{Mod}}(F M, X) \xrightarrow{\simeq} \operatorname{Mor}_{A-\operatorname{Mod}}(M, G X) .
$$

Let $L \stackrel{f}{\rightarrow} M \xrightarrow{g} N \rightarrow 0$ an exact sequence of $A$-modules. The naturality of the map $\theta$ implies that the following diagram commutes for any $X \in B$-Mod:


Since $\operatorname{Mor}_{A-\operatorname{Mod}(-, G X)}$ is left-exact then the inferior row is exact, and since the vertical morphisms are isomorphisms, then the superior row is also exact. Using Proposition 10.3.32 item (b) we obtain that $F L \xrightarrow{F f} F M \xrightarrow{F g} F N \rightarrow 0$, therefore $F$ is right exact. Analogously we can show that $G$ is left exact.

### 10.4 Bimodules

Definition 10.4.1 (Bimodule). Let $R$ and $S$ be any two rings, and $M$ be an abelian group. We say that $M$ is a $(R, S)$-bimodule if $M$ is both a left- $R$-module and a right- $S$-module, and $M$ must satisfy the associative law:

$$
r(m s)=(r m) s,
$$

for all $r \in R, m \in M$, and $s \in S$.
Example 10.4.2. Let $R$ be a commutative ring and $M$ be an $R$-module, then we can define the structure of $(R, R)$-bimodule on $M$ by making the identification $r m=m r$ for any $r \in R$ and $M$. Such identification respects the module structure since, given any other $r^{\prime} \in R$ one has

$$
\left(r r^{\prime}\right) m=r\left(r^{\prime} m\right)=r^{\prime} m r=(m r) r^{\prime}=m\left(r r^{\prime}\right)=m\left(r^{\prime} r\right)
$$

Proposition 10.4.3. Let $R$ and $S$ be rings, and $M$ be a ( $R, S$ )-bimodule. The following properties concern the relations between functors $\operatorname{Mor}(M,-)$ and $\operatorname{Mor}(-, M)$, and bimodules:
(a) For any left-R-module $B$ we have that $\operatorname{Mor}_{R} \operatorname{Mod}(M, B)$ is a left-S-module with a product $s \cdot f: M \rightarrow B$ mapping $m \mapsto f(m s)$ for any $f \in \operatorname{Mor}_{R} \operatorname{Mod}(M, B)$ and $s \in S$. This can be summarized by the fact that

$$
\operatorname{Mor}_{R} \operatorname{Mod}(M,-):{ }_{R} \operatorname{Mod} \longrightarrow{ }_{S} \operatorname{Mod}
$$

is a covariant functor.
(b) For any right-S-module $B$ we have that $\operatorname{Mor}_{\text {Mod }_{S}}(M, B)$ is a right- $R$-module with a product $f \cdot r: M \rightarrow B$ mapping $m \mapsto f(r m)$ for any $f \in \operatorname{Mor}_{\text {Mod }_{S}}(M, B)$ and $s \in S$. This can be summarized by the fact that

$$
\operatorname{Mor}_{\text {Mod }_{S}}(M,-): \operatorname{Mod}_{S} \longrightarrow \operatorname{Mod}_{R}
$$

is a covariant functor.
(c) For any right-S-module $A$ we have that $\operatorname{Mor}_{\text {Mod }_{S}}(A, M)$ is a left- $R$-module with a product $r \cdot f: A \rightarrow M$ mapping $m \mapsto r f(m)$. This is summarized by the fact that

$$
\operatorname{Mor}_{\text {Mod }_{S}}(-, M): \operatorname{Mod}_{S}{ }^{\mathrm{op}} \longrightarrow{ }_{R} \operatorname{Mod}
$$

is a contravariant functor.
(d) For any left-R-module $A$ we have that $\operatorname{Mor}_{R} \operatorname{Mod}(A, M)$ is a right-S-module with a product $f \cdot s: A \rightarrow M$ mapping $m \mapsto f(m) s$. This is summarized by the fact that

$$
\operatorname{Mor}_{R} \operatorname{Mod}(-, M):{ }_{R} \operatorname{Mod}^{\mathrm{op}} \longrightarrow \operatorname{Mod}_{S}
$$

is a contravariant functor.
Proof. The proof of the above statements are rather repetitive, but we shall lay all of them down:
(a) Given a morphism $f \in \operatorname{Mor}_{R} \operatorname{Mod}(M, B)$, elements $s, s^{\prime} \in S$ and any $m \in M$, one has

$$
\left(\left(s s^{\prime}\right) \cdot f\right)(m)=f\left(m\left(s s^{\prime}\right)\right)=f\left((m s) s^{\prime}\right)=\left(s^{\prime} \cdot f\right)(m s)=\left(s \cdot\left(s^{\prime} \cdot f\right)\right)(m)
$$

therefore $\left(s s^{\prime}\right) \cdot f=s \cdot\left(s^{\prime} \cdot f\right)$. This shows that $\operatorname{Mor}_{R} \operatorname{Mod}(M, B)$ is a left- $S$-module. Now, if $g: B \rightarrow C$ is any morphism of left- $R$-modules, consider the map $g_{*}: \operatorname{Mor}_{R} \operatorname{Mod}(M, B) \rightarrow$ $\operatorname{Mor}_{R} \operatorname{Mod}(M, C)$. Take a morphism $f \in \operatorname{Mor}_{R} \operatorname{Mod}(M, B)$ and consider any two elements $s \in S$ and $m \in M$, then

$$
g_{*}(s f)(m)=g(s f)(m)=g f(m s)=s(g f(m))=s g_{*}(f)(m),
$$

which shows that $g_{*}$ is indeed a morphism between left-S-modules.
(b) Let $f \in \operatorname{Mor}_{\text {Mod }_{S}}(M, B)$, and $r, r^{\prime} \in R$ be any elements, then given any $m \in M$ one has

$$
\left(f \cdot\left(r r^{\prime}\right)\right)(m)=f\left(\left(r r^{\prime}\right) m\right)=f\left(r\left(r^{\prime} m\right)\right)=(f \cdot r)\left(r^{\prime} m\right)=\left((f \cdot r) \cdot r^{\prime}\right)(m)
$$

Therefore $\operatorname{Mor}_{\text {Mod }_{s}}(M, B)$ has a structure of right- $R$-module. Let $g: B \rightarrow C$ be any morphism of right-S-modules and consider its induced map $g_{*}: \operatorname{Mor}_{\text {Mods }}(M, B) \rightarrow$ $\operatorname{Mor}_{\text {Mod }_{S}}(M, C)$. If $f \in \operatorname{Mor}_{\text {Mod }_{S}}(M, B)$ is any morphism, consider elements $r \in R$ and $m \in M$, then

$$
g_{*}((f \cdot r))(m)=g(f \cdot r)(m)=g f(r m)=r(g f(m))=r g_{*}(f)(m)
$$

Therefore $g_{*}$ is a morphism between right- $R$-modules.
(c) Let $f \in \operatorname{Mor}_{M_{0 d} S}(A, M)$ be any morphism, and consider two elements $r, r^{\prime} \in R$. From the product definition we obtain, for any $m \in M$ :

$$
\left(\left(r r^{\prime}\right) \cdot f\right)(m)=\left(r r^{\prime}\right) f(m)=r\left(r^{\prime} f(m)\right)=r\left(r^{\prime} \cdot f\right)(m)=\left(r \cdot\left(r^{\prime} \cdot f\right)\right)(m)
$$

which shows that $\operatorname{Mor}_{\text {Mod }_{S}}(A, M)$ has a structure of left- $R$-module. For the second part of the proposition, consider a morphism $g: A \rightarrow B$ between right- $S$ modules, and its corresponding map $g^{*}: \operatorname{Mor}_{\text {Mod }_{S}}(B, M) \rightarrow \operatorname{Mor}_{\text {Mod }_{S}}(A, M)$. Let $f \in \operatorname{Mor}_{\text {Mod }_{S}}(B, M), r \in R$, and $b \in B$ be any elements, then one has

$$
g^{*}(r \cdot f)(m)=(r \cdot f) \circ g(m)=r f(g(m))=r g^{*}(f)(m)
$$

Therefore $g^{*}$ is a morphism between left- $R$-modules.
(d) Let $f \in \operatorname{Mor}_{R} \operatorname{Mod}(A, M)$ be any morphism and take elements $s, s^{\prime} \in S$, then for any $m \in M$ we have

$$
\left(f \cdot\left(s s^{\prime}\right)\right)(m)=f(m)\left(s s^{\prime}\right)=(f(m) s) s^{\prime}=(f \cdot s)(m) s^{\prime}=\left((f \cdot s) \cdot s^{\prime}\right)(m)
$$

therefore $\operatorname{Mor}_{R} \operatorname{Mod}(A, M)$ has the structure of a right-S-module. For the last part, take any morphism $g: A \rightarrow B$ of left- $R$-modules and consider the induced map $g^{*}: \operatorname{Mor}_{R} \operatorname{Mod}(B, M) \rightarrow \operatorname{Mor}_{R} \operatorname{Mod}(A, M)$. If $f \in \operatorname{Mor}_{R} \operatorname{Mod}(B, M), s \in S$, and $b \in B$ are any elements, then one has

$$
g^{*}(f \cdot s)(m)=(f \cdot s) \circ g(m)=f g(m) s g^{*}(f)(m) s .
$$

This shows that $g^{*}$ is a morphism between right- $S$-modules.

Example 10.4.4. If $R$ is a commutative ring, and $A$ and $B$ are $R$-modules, then both modules also have the structure of $(R, R)$-bimodules (see Example 10.4.2). From Proposition 10.4.3 we find that

$$
\begin{aligned}
& \operatorname{Mor}_{(R, R)-\operatorname{Mod}}(A,-):(R, R)-\operatorname{Mod} \longrightarrow(R, R)-\operatorname{Mod} \\
& \operatorname{Mor}_{(R, R)-\operatorname{Mod}}(-, B):(R, R)-\operatorname{Mod}^{\mathrm{op}} \longrightarrow(R, R)-\operatorname{Mod}
\end{aligned}
$$

are the Mor functors.
Corollary 10.4.5. Let $R$ be a ring and $M$ be a left- $R$-module. Then $\operatorname{Mor}_{R} \operatorname{Mod}(R, M)$ is a left- $R$-module and there exists a natural isomorphism of left- $R$-modules

$$
\operatorname{Mor}_{R} \operatorname{Mod}(R, M) \simeq M,
$$

mapping $f \mapsto f(1)$.
Proof. Since $R$ has a natural structure of $(R, R)$-bimodule, by means of item (a) of Proposition 10.4.3 we obtain that $\operatorname{Mor}_{R} \operatorname{Mod}(R, M)$ has a structure of left- $R$-module via the product $r \cdot f \in \operatorname{Mor}_{R} \operatorname{Mod}(R, M)$ mapping $a \mapsto f(a r)$ —where $r \in R$ and $f \in \operatorname{Mor}_{R} \operatorname{Mod}(R, M)$ are any two elements.

We first shows that the map $\phi: \operatorname{Mor}_{R} \operatorname{Mod}(R, M) \rightarrow M$ given by $f \mapsto f(1)$ is a morphism of left- $R$-modules. Let $f, g \in \operatorname{Mor}_{R} \operatorname{Mod}(R, M)$ be any two elements, then $\phi(f+g)=(f+g)(1)=f(1)+g(1)=\phi(f)+\phi(g)$. Moreover, if $r \in R$ then

$$
\phi(r \cdot f)=(r \cdot f)(1)=f(1 \cdot r)=f(r)=f(r \cdot 1)=r f(1)=r \phi(f) .
$$

Therefore $\phi$ is indeed a morphism as wanted. For injectivity it is simple to realize that $\phi(f)=f(1)=0$ if and only if $f=0$, since $f(r)=f(r \cdot 1)=r f(1)$-therefore ker $\phi=0$. Moreover, given any $m \in M$, there exists a morphism $g: R \rightarrow M$ uniquely determined by $g(1):=m$, so that $\phi(g)=m$.

Theorem 10.4.6. Let $R$ and $S$ be any two rings, and $M$ be an $(R, S)$-bimodule. Then for any collection $\left(B_{j}\right)_{j \in J}$ of left-R-modules, the natural isomorphism of abelian groups

$$
\operatorname{Mor}_{R} \operatorname{Mod}\left(A, \prod_{j \in J} B_{j}\right) \simeq \prod_{j \in J} \operatorname{Mor}_{R} \operatorname{Mod}\left(A, B_{j}\right)
$$

is also an isomorphism of left-S-modules.
Proof. Define $B:=\prod_{j \in J} B_{j}$, and name the isomorphism by $\phi$-which maps $f \mapsto$ $\left(\pi_{j} f\right)_{j \in J}$, where $\pi_{j}$ is the $j$-th canonical projection. Notice that, given any morphism $f \in \operatorname{Mor}_{R} \operatorname{Mod}(A, B), s \in S$ and $a \in A$, we have

$$
\pi_{j} \circ(s \cdot f)(a)=\pi_{j}(f(a s))=\pi_{j} f(a s)=\left(s \cdot\left(\pi_{j} f\right)\right)(a)
$$

for any $j \in J$. Therefore, we conclude that

$$
\phi(s \cdot f)=\left(\pi_{j}(s \cdot f)\right)_{j \in J}=\left(s \cdot\left(\pi_{j} f\right)\right)_{j \in J}=s\left(\pi_{j} f\right)_{j \in J}=s \phi(f),
$$

which proves that $\phi$ is a morphism of left-S-modules. Since $\phi$ is also an isomorphism of abelian groups, then $\phi$ is bijective, thus $\phi$ is indeed an isomorphism of left-Smodules.

Definition 10.4.7 (Dual module). Let $M$ be a right- $R$-module (or left). We define the dual module of $M$ to be the left- $R$-module (or right):

$$
M^{*}:=\operatorname{Mor}_{\operatorname{Mod}_{R}}(M, R)
$$

### 10.5 Projective Modules

## Lifting Property For Projective Modules

Theorem 10.5.1 (Free modules have the lifting property). Let $R$ be a ring and $F$ be a free $R$-module. For every surjective morphism of $R$-modules $p: M \rightarrow N$ and morphism $h: F \rightarrow N$, there exists a unique morphism $\ell: F \rightarrow M$, called lifting of $h$, such that the diagram

commutes in $R$-Mod.
Proof. Since $F$ is free, let $B:=\left(b_{j}\right)_{j \in J}$ be a basis for $F$. From the surjectivity of $p$, for every $j \in J$ there exists $m_{j} \in M$ such that $p\left(m_{j}\right)=h\left(b_{j}\right)$. From the free module universal property there exists a unique $\ell: F \rightarrow M$ such that $\ell\left(b_{j}\right)=m_{j}$ for each $j \in J$. From construction we have $p \ell\left(b_{j}\right)=p\left(m_{j}\right)=h\left(b_{j}\right)$, therefore $p \ell=h$, since $B$ generates $F$, and the diagram commutes.

Remark 10.5.2 (Uniqueness of the lift). The lift $\ell$ of $h$ need not be unique in case $F$ isn't free!

Definition 10.5.3 (Projective module). Let $R$ be a ring and $P$ be an $R$-module. We say that $P$ is a projective $R$-module if

$$
\operatorname{Mor}_{R-\operatorname{Mod}}(P,-): R-\operatorname{Mod} \longrightarrow \mathrm{Ab}
$$

is an exact covariant functor.
Proposition 10.5.4 (Equivalences for projective modules). Let $R$ be a ring and $P$ be an $R$-module. The following properties are equivalent:
(a) The module $P$ is projective.
(b) For every exact sequence of $R$-modules $M \stackrel{g}{\rightarrow} N \rightarrow 0$ the sequence of abelian groups

$$
\operatorname{Mor}_{R-\operatorname{Mod}}(P, M) \xrightarrow{g_{*}} \operatorname{Mor}_{R-\operatorname{Mod}}(P, N) \longrightarrow 0
$$

is exact. Equivalently, for every $h: P \rightarrow N$, there exists a lifting $\ell: P \rightarrow N$ of $h-n o t$ necessarily unique-such that the diagram

is commutative in $R$-Mod.
(c) Every short exact sequence of $R$-modules of the form

$$
0 \longrightarrow L \longmapsto M \longrightarrow P \longrightarrow 0
$$

is a split sequence.
(d) The module $P$ is a direct summand of a free $R$-module.
(e) There exists a collection $\left(x_{j}\right)_{j \in J}$ of elements $x_{j} \in P$, and a collection of associated morphisms of $R$-modules $\left(\phi_{j}: P \rightarrow R\right)_{j \in J}^{3}$ such that, for all $x \in P$ we have:

- The elements $\phi_{j}(x) \in R$ are non-zero for only finitely many $j \in J$.
- The element $x$ can be written as $x=\sum_{j \in J} x_{j} \phi_{j}(x)$.

Proof. From the definition, the equivalence of (a) and (b) is immediate. We prove the following:

[^16]- (b) $\Rightarrow$ (c). Let $g: M \rightarrow P$ be the epimorphism depicted in the sequence. Consider the identity morphism $\operatorname{id}_{P}: P \rightarrow P$ and apply (b) to obtain $\rho: P \rightarrow M$ such that

is a commutative diagram of $R$-modules. Notice that $g \rho=\operatorname{id}_{p}$, therefore $\rho$ is a section of $g$-thus the sequence splits.
- $(\mathrm{c}) \Rightarrow(\mathrm{d})$. Via Theorem 10.2.20 let $p: F \rightarrow M$ be a surjective morphism of $R$ modules, where $F$ is free. Thus we have a short exact sequence

$$
0 \longrightarrow \operatorname{ker} p \longrightarrow F \xrightarrow{p} P \longrightarrow 0 .
$$

By item (c) we find that the above sequence is split, therefore if $\iota: P \mapsto F$ is a section of $p$, then

$$
F=\operatorname{ker} p \oplus \operatorname{im} \iota \simeq \operatorname{ker} p \oplus P
$$

- $(\mathrm{d}) \Rightarrow(\mathrm{b})$. Let $f: M \rightarrow N$ be a surjective morphism of $R$-modules, and $\psi: P \rightarrow N$ be any morphism. By item (d), let $F$ be a free module with $F \simeq P \oplus P^{\prime}$, where $P^{\prime}$ is the complement $R$-module of $P$ with respect to $F$-also, let $B$ be a basis of $F$. Considering the natural projection $\pi_{P}: F \rightarrow P$, define a map $\phi^{\prime}: F \rightarrow M$ as follows: given $b \in B$, by the surjectivity of $f$, there exists $m \in M$ such that $f(m)=\psi \pi_{P}(b)$ —we shall define $\phi^{\prime}(b):=m$. It is easily seen that $\phi^{\prime}$ is $R$-linear, and that $f \phi^{\prime}=\psi \pi_{p}$. Moreover, the surjectivity of $f$ implies that $\phi^{\prime}$ is the unique morphism of modules with such property. Considering the natural inclusion $\iota_{P}: P \hookrightarrow F$ which is a section of $\pi_{P}$ —define $\phi:=\phi^{\prime} \iota_{P}: P \rightarrow M$ and notice that

$$
f \phi=f\left(\phi^{\prime} \iota_{P}\right)=\left(f \phi^{\prime}\right) \iota_{P}=\left(\psi \pi_{P}\right) \iota_{P}=\psi\left(\pi_{P} \iota_{P}\right)=\psi .
$$

This finishes the proof of the equivalence of the items (a), (b), (c), and (d). For item (e), we shall prove its equivalence with (d).

- (d) $\Rightarrow$ (e). Via item (d), there exists an indexing set $J$ and an isomorphism $\psi: \bigoplus_{j \in J} R \xlongequal{\leftrightharpoons} P \oplus P^{\prime}$, where $P^{\prime}$ is the complement of $P$. If $\pi_{P}: P \oplus P^{\prime} \rightarrow P$ denotes the canonical projection, define a collection $\left(x_{j}\right)_{j \in J}$ by $x_{j}:=\pi_{P} \psi\left(e_{j}\right)$-where $e_{j}:=$ $\left(\delta_{i j}\right)_{i \in J}$. Let $\iota_{P}: P \hookrightarrow P \oplus P^{\prime}$ be the canonical inclusion of $P$, and $\pi_{j}: \bigoplus_{j \in J} R \rightarrow R$ be the canonical projection of the $j$-th coordinate. Define a collection of morphisms $\left(\phi_{j}\right)_{j \in J}$ by $\phi_{j}:=\pi_{j} \psi^{-1} \iota_{P}: P \rightarrow R$, so that-since $\psi^{-1} \iota_{P}(x) \in \bigoplus_{j \in J} R$ has finitely many non-zero components-there are finitely many $j \in J$ such that $\pi_{j} \psi^{-1} \iota_{P}(x) \in$
$R$ is non-zero. For the last condition of item (e), if $x \in P$ is any element, we have

$$
\begin{aligned}
x & =\pi_{P} \iota_{P}(x)=\pi_{P}\left(\psi \psi^{-1}\right) \iota_{P}(x)=\pi_{P} \psi\left(\psi^{-1} \iota_{P}\right)(x)=\pi_{P} \psi\left(\phi_{j}(x)\right)_{j \in J} \\
& =\pi_{P} \psi\left(\sum_{j \in J} e_{j} \phi_{j}(x)\right)=\sum_{j \in J} \pi_{P} \psi\left(e_{j} \phi_{j}(x)\right)=\sum_{j \in J} \pi_{P} \psi\left(e_{j}\right) \phi_{j}(x) \\
& =\sum_{j \in J} x_{j} \phi_{j}(x) .
\end{aligned}
$$

- $(\mathrm{e}) \Rightarrow(\mathrm{d})$. The collection $\left(\phi_{j}\right)_{j \in J}$ induces, by the universal property of the product, a unique morphism of $R$-modules $\phi: P \rightarrow \prod_{j \in J} R$ mapping $x \mapsto\left(\phi_{j}(x)\right)_{j \in J}$. For any $x \in P$ we know from hypothesis that the collection $\left(\phi_{j}(x)\right)_{j \in J}$ has finitely many non-zero elements, therefore $\operatorname{im} \phi \subseteq \bigoplus_{j \in J} R$. Then we may naturally restrict the codomain of $\phi$, obtaining a morphism $\phi: P \rightarrow \bigoplus_{j \in J} R$. Define a collection $\left(e_{j}\right)_{j \in J}$ where $e_{j}:=\left(\delta_{i j}\right)_{i \in J}$, and consider an $R$-linear map $\lambda: \bigoplus_{j \in J} R \rightarrow P$ defined by sending $e_{j} \mapsto x_{j}$. By hypothesis, one has

$$
\lambda \phi(x)=\lambda\left(\phi_{j}(x)\right)_{j \in J}=\sum_{j \in J} x_{j} \phi_{j}(x)=x
$$

therefore $\phi$ is a section of $\lambda$, showing that $\lambda$ is a split epimorphism. Thus $P$ is a direct summand of the free module $\bigoplus_{j \in J} R$.

Example 10.5.5. From Theorem 10.5 .1 we find that every free module is a projective module.

Example 10.5.6. Let $R$ be a ring. If there exists an idempotent element $e \in R$ (that is, $\left.e^{2}=e\right)$, then we have a decomposition $R=e R \oplus(1-e) R$. Therefore $e R$ is projective.

Indeed, given $r \in R$ we can write it as $r=e r+(1-e) r$-so that $R=e R+(1-e) R$. Moreover, if $a \in e R \cap(1-e) R$, let $r, r^{\prime} \in R$ be such that $a=e r=(1-e) r^{\prime}$, however, notice that

$$
a=e r=e(e r)=e\left((1-e) r^{\prime}\right)=\left(e-e^{2}\right) r^{\prime}=(e-e) r^{\prime}=0
$$

Therefore $e R \cap(1-e) R=0$, thus $R=e R \oplus(1-e) R$.
As an example of a projective module that isn't free: if $e$ is a central in $R$, then given any er $\in e R$ we have

$$
(e r)(1-e)=e r-(e r) e=e r-e(r e)=e r-e(e r)=e r-e r=0
$$

That is, the singleton $\{e r\}$ is $R$-linearly dependent-therefore $e R$ does not admit a basis.
Proposition 10.5.7. Let $\left(P_{j}\right)_{j \in J}$ be a family of $R$-modules. Then the module $P:=\bigoplus_{j \in J} P_{j}$ is projective if and only if $P_{j}$ is projective for each $j \in J$.

Proof. Suppose that $P$ is projective, then we let $F$ be a free module of which $P$ is a direct summand, say $F=P \oplus P^{\prime}$. If $\sigma: J \rightarrow J$ is any permutation, we know that $\bigoplus_{j \in J} P_{j} \simeq \bigoplus_{j \in J} P_{\sigma(j)}$, therefore for all $i \in J$ we have

$$
F=P \oplus P^{\prime}=\left(\bigoplus_{j \in J} P_{j}\right) \oplus P^{\prime} \simeq\left(P_{i} \oplus \bigoplus_{j \in J \backslash i} P_{j}\right) \oplus P^{\prime}=P_{i} \oplus\left(\left(\bigoplus_{j \in J \backslash i} P_{j}\right) \oplus P^{\prime}\right)
$$

Since $P_{i}$ is a direct summand of a free module, it is a projective module.
For the converse, suppose that $P_{j}$ is projective for all $j \in J$. Let $g: M \rightarrow N$ be a surjective morphism of $R$-modules, and $\phi: P \rightarrow N$ be any morphism. Since $P_{j}$ is projective, if $\iota_{j}: P_{j} \hookrightarrow P$ is the canonical inclusion, there exists a morphism $\psi_{j}: P_{j} \rightarrow M$ making the diagram

commute in $R$-Mod. By the universal property of the coproduct $P$, the collection of morphisms $\left(\psi_{j}\right)_{j \in J}$ induce a unique morphism $\psi: P \rightarrow M$ such that $\psi \iota_{j}=\psi_{j}$ for all $j \in J$. To show that $g \psi=\phi$, notice that, given any $\left(x_{j}\right)_{j \in J} \in P$, one has

$$
g \psi\left(x_{j}\right)_{j \in J}=g\left(\sum_{j \in J} \psi_{j}\left(x_{j}\right)\right)=\sum_{j \in J} g \psi_{j}\left(x_{j}\right)=\sum_{j \in J} \phi \iota_{j}\left(x_{j}\right)=\phi\left(x_{j}\right)_{j \in J}
$$

Therefore, the following diagram commutes in $R$-Mod:

which shows that $P$ is a projective module.
Proposition 10.5.8 (Eilenberg's trick). Let $P$ be a projective $R$-module. Then there exists a free $R$-module $F$ such that the direct sum $P \oplus F$ is a free module.

Proof. Since $P$ is projective, there exists a free $R$-module $F^{\prime}$ and an $R$-module $Q$ such that $F^{\prime}=P \oplus Q$. Notice that

$$
(P \oplus Q) \oplus F^{\prime}=P \oplus(Q \oplus(P \oplus Q))=P \oplus((Q \oplus P) \oplus Q)
$$

is a direct sum of free modules, thus also free-however, $(Q \oplus P) \oplus Q$ may not be free, since $Q$ is only ensured to be projective. However, if we define $F:=\bigoplus_{j \in \mathbf{N}}(Q \oplus P)$, then $F$ is a direct sum of projective modules and hence projective itself. Using the fact that $P \oplus(Q \oplus P)=(P \oplus Q) \oplus P$, we obtain

$$
P \oplus F=P \oplus \bigoplus_{j \in \mathbf{N}}(Q \oplus P)=\bigoplus_{j \in \mathbf{N}}(P \oplus Q)
$$

which is a direct sum of free modules-hence a projective module. Therefore $F$ satisfies the requirement of the statement.

Proposition 10.5.9 (Schanuel). Let $P$ and $P^{\prime}$ be projective $R$-modules. If there exists short exact sequences

$$
\begin{aligned}
& 0 \longrightarrow K \xrightarrow{f} P \xrightarrow{g} M \longrightarrow 0 \\
& 0 \longrightarrow K^{\prime} \not{\longleftrightarrow} P^{\prime} \xrightarrow{\psi} M \longrightarrow 0
\end{aligned}
$$

of $R$-modules, then there exists an isomorphism of $R$-modules

$$
K \oplus P^{\prime} \simeq K^{\prime} \oplus P
$$

Proof. Since $P$ is projective, there exists a morphism of $R$-modules $\varepsilon: P \rightarrow P^{\prime}$ such that the diagram

commutes in $R$-Mod. Notice that for any $k \in K$ one has

$$
\psi \varepsilon f(k)=g f(k)=0,
$$

therefore $\varepsilon f(k) \in \operatorname{ker} \psi$. Since $\operatorname{ker} \psi \subseteq \operatorname{im} \phi$, there must exist $k^{\prime} \in K^{\prime}$ such that $\phi\left(k^{\prime}\right)=\varepsilon f(k)$-which needs to be unique with such image since $\phi$ is injective. Define $\delta: K \rightarrow K^{\prime}$ to be the map $k \mapsto k^{\prime}$, where $k^{\prime} \in K^{\prime}$ is as described above. So far, we have the following commutative diagram in $R$-Mod


We now define a map $\omega: K \rightarrow K^{\prime} \oplus P$ given by $k \mapsto(\delta(k), f(k))$, and $\gamma: K^{\prime} \oplus P \rightarrow P^{\prime}$ by $\left(k^{\prime}, p\right) \mapsto \varepsilon(p)-\phi\left(k^{\prime}\right)$. Notice that both maps are clearly $R$-linear and from definition $\left(k^{\prime}, p\right) \in \operatorname{ker} \gamma$, that is, $\varepsilon(p)=\phi\left(k^{\prime}\right)$ if and only if there exists a common $k \in K$ for which $f(k)=p$ and $\delta(k)=k^{\prime}$-therefore $\operatorname{ker} \gamma=\operatorname{im} \omega$. To show that $\omega$ is injective, let $k \in \operatorname{ker} \omega$ be any element, then from definition $(\delta(k), f(k))=(0,0)$, but since $f$ is injective, then $k=0$-thus $\operatorname{ker} \omega=0$ and $\omega$ is injective. For the surjectivity of $\gamma$, if $p^{\prime} \in P^{\prime}$, let $p \in P$ be such that $g(p)=\psi\left(p^{\prime}\right)$-which exists because $g$ is surjective. Then by the commutativity of the right square one has

$$
\psi \varepsilon(p)=g(p)=\psi\left(p^{\prime}\right)
$$

that is, $\varepsilon(p)-p^{\prime} \in \operatorname{ker} \psi$ and thus exists $k^{\prime} \in K^{\prime}$ for which $\phi\left(k^{\prime}\right)=\varepsilon(p)-p^{\prime}$. This shows that

$$
\gamma\left(k^{\prime}, p^{\prime}\right)=\varepsilon(p)-\phi\left(k^{\prime}\right)=p^{\prime}
$$

so that $\gamma$ is surjective.
We've just shown that the sequence

$$
0 \longrightarrow K \xrightarrow{\omega} K^{\prime} \oplus P \xrightarrow{\gamma} P^{\prime} \longrightarrow 0
$$

is short exact. Since $P^{\prime}$ is projective, the sequence is also split and therefore there exists an isomorphism of $R$-modules $K^{\prime} \oplus P \simeq K \oplus P^{\prime}$.

Theorem 10.5.10 (Projective resolution). For any given $R$-module $M$ there exists an exact sequence

$$
\cdots \longrightarrow P_{n} \xrightarrow{f_{n}} P_{n-1} \xrightarrow{f_{n-1}} \cdots \longrightarrow P_{1} \xrightarrow{f_{1}} P_{0} \xrightarrow{f_{0}} M \longrightarrow 0
$$

where each $R$-module $P_{j}$ is projective.
Proof. Notice that since every $R$-module is the quotient of a free $R$-module (see Theorem 10.2.20), which is projective, then $M \simeq P_{0} / T$ for some free $R$-module $P_{0}$. Take $f_{0}: P_{0} \rightarrow M$ to be the canonical projection. For the remainder of the sequence take $\left(P_{j}, f_{j}\right)=\operatorname{ker} f_{j-1}$ for $j>0$.

## Finitely Generated Projective Modules

Proposition 10.5.11. Let $P$ be a finitely generated projective right- $R$-module (or left). Then the dual module $P^{*}$ is a projective left- $R$-module (or right).

Proof. If $P$ is finitely generated and projective we can choose a finitely generated free right- $R$-module $F$ of which $P$ is a direct summand-say, $F=P \oplus P^{\prime}$. Since $F$ is finitely generated, there exists $n \in \mathbf{Z}_{>0}$ such that $F \simeq \bigoplus_{j=1}^{n} R$. We know that there exists a natural isomorphism of left-R-modules

$$
F^{*}=\operatorname{Mor}_{\operatorname{Mod}_{R}}\left(P \oplus P^{\prime}, R\right) \simeq \operatorname{Mor}_{\operatorname{Mod}_{R}}(P, R) \oplus \operatorname{Mor}_{\operatorname{Mod}_{R}}\left(P^{\prime}, R\right)=P^{*} \oplus P^{\prime *} .
$$

Moreover, we can rewrite the morphism set using the fact that $F$ is finitely generated:

$$
F^{*} \simeq \operatorname{Mor}_{\operatorname{Mod}_{R}}\left(\bigoplus_{j=1}^{n} R, R\right) \simeq \bigoplus_{j=1}^{n} \operatorname{Mor}_{\operatorname{Mod}_{R}}(R, R) \simeq \bigoplus_{j=1}^{n} R,
$$

therefore $F^{*}$ is a finitely generated free left- $R$-module. Since $P^{*}$ is a direct summand of $F^{*}$, then $P^{*}$ is a finitely generated projective left- $R$-module.

Proposition 10.5.12. Let $P$ be a right- $R$-module (or left). The double dual $P^{* *}$ is a right-$R$-module (or left), and the natural evaluation map

$$
\mathrm{eval}: P \longrightarrow P^{* *}
$$

sending $x \mapsto \operatorname{eval}_{x}: P^{*} \rightarrow R$ —where $\operatorname{eval}_{x}(f)=f(x)$-is an injective morphism of right-$R$-modules.

Proof. Let $\left(x_{j}, \phi_{j}\right)_{j \in J}$ be the dual "basis" of $P$, where $x_{j} \in P$ and $\phi_{j} \in P^{*}$. If $x \in$ ker eval, then $f(x)=0$ for all $f \in P^{*}$-thus in particular $x=\sum_{j \in J} x_{j} \phi_{j}(x)=0$. This shows that ker eval $=0$, thus the evaluation morphism is injective.

Proposition 10.5.13. Let $P$ be a finitely generated projective right- $R$-module (or left), with a dual "basis" $\left(x_{j}, \phi_{j}\right)_{j=1}^{n}$. The following properties hold:
(a) The finite collection $\left(\phi_{j}, \text { eval }_{x_{j}}\right)_{j=1}^{n}$ forms a dual "basis" for the dual module $P^{*}$.
(b) The dual $P^{*}$ is a finitely generated projective left- $R$-module (or right), with a generating set $\left(\phi_{j}\right)_{j=1}^{n}$.
(c) The double dual $P^{* *}$ is a finitely generated projective right- $R$-module (or left), with a generating set $\left(\text { eval }_{x_{j}}\right)_{j=1}^{n}$.
(d) The natural evaluation morphism eval: $P \mapsto P^{* *}$ is an isomorphism of right- $R$ modules (or left).

Proof. (a) Let $f \in P^{*}$ be any functional, then given any $x \in P$ one has

$$
\sum_{j=1}^{n} \operatorname{eval}_{x_{j}}(f) \phi_{j}(x)=\sum_{j=1}^{n} f\left(x_{j} \phi_{j}(x)\right)=f\left(\sum_{j=1}^{n} x_{j} \phi_{j}(x)\right)=f(x)
$$

therefore $f=\sum_{j=1}^{n} \operatorname{eval}_{x_{j}}(f) \phi_{j}$-which proves that $\left(\phi_{j}, \mathrm{eval}_{x_{j}}\right)_{j}$ is a dual basis for $P^{*}$.
(b) It is immediate from the last item's proof that $\left(\phi_{j}\right)_{j=1}^{n}$ is a generating set for $P^{*}$. Moreover, since $P^{*}$ admits a dual basis it follows that $P^{*}$ is projective.
(c) Define a map eval ${ }^{*}: P^{*} \rightarrow\left(P^{* *}\right)^{*}$ given by $f \mapsto \operatorname{eval}_{f}^{*} \in \operatorname{Mor}_{\operatorname{Mod}_{R}}\left(P^{* *}, R\right)$, where $\operatorname{eval}_{f}^{*}(\Phi):=\Phi(f)$ for any $\Phi \in \operatorname{Mor}_{R} \operatorname{Mod}\left(P^{*}, R\right)$. Notice that, for any $f \in P^{*}$ one has
$\sum_{j=1}^{n} \operatorname{eval}_{x_{j}}(f) \operatorname{eval}_{\phi_{j}}^{*}(\Phi)=\sum_{j=1}^{n} f\left(x_{j}\right) \Phi\left(\phi_{j}\right)=\sum_{j=1}^{n} \Phi\left(f\left(x_{j}\right) \phi_{j}\right)=\Phi\left(\sum_{j=1}^{n} f\left(x_{j}\right) \phi_{j}\right)=\Phi(f)$,
that is, $\Phi=\sum_{j=1}^{n} \operatorname{eval}_{x_{j}} \cdot \operatorname{eval}_{\phi_{j}}^{*}(\Phi)$. Therefore $\left(\operatorname{eval}_{x_{j}}\right)_{j=1}^{n}$ is a generating set for $P^{* *}$, and (eval $\left.{ }_{\phi_{j}}^{*}, \mathrm{eval}_{x_{j}}\right)_{j=1}^{n}$ is a dual "basis" for $P^{* *}$. Therefore $P^{* *}$ is a finitely generated projective right- $R$-module.
(d) Given an element $\Phi \in P^{* *}$, we can rewrite it as $\Phi=\sum_{j=1}^{n} \operatorname{eval}_{x_{j}} \cdot$ eval $_{\phi_{j}}^{*}(\Phi)$, therefore, if we take $\sum_{j=1}^{n} x_{j} \Phi\left(\phi_{j}\right) \in P$, one obtains

$$
\begin{aligned}
\operatorname{eval}\left(\sum_{j=1}^{n} x_{j} \Phi\left(\phi_{j}\right)\right) & =\sum_{j=1}^{n} \operatorname{eval}\left(x_{j} \Phi\left(\phi_{j}\right)\right)=\sum_{j=1}^{n} \operatorname{eval}\left(x_{j}\right) \Phi\left(\phi_{j}\right) \\
& =\sum_{j=1}^{n} \operatorname{eval}_{x_{j}} \cdot \Phi\left(\phi_{j}\right)=\sum_{j=1}^{n} \operatorname{eval}_{x_{j}} \cdot \operatorname{eval}_{\phi_{j}}^{*}(\Phi) \\
& =\Phi
\end{aligned}
$$

This shows that eval is also a surjective morphism. Therefore eval is an isomorphism of right- $R$-modules

$$
P \simeq P^{* *} .
$$

### 10.6 Injective Modules

## Lifting Property for Injective Modules

Definition 10.6.1 (Injective module). Let $R$ be a ring, and $E$ be an $R$-module. We say that $E$ is an injective $R$-module if

$$
\operatorname{Mor}_{R-M o d}(-, E): R-\operatorname{Mod}^{\mathrm{op}} \longrightarrow \mathrm{Ab}
$$

is an exact contravariant functor.
Proposition 10.6.2. Let $E$ be an $R$-module. The following are equivalent properties:
(a) The module $E$ is injective.
(b) Given any exact sequence of $R$-modules $0 \rightarrow M \stackrel{f}{\mapsto} N$, the sequence

$$
\operatorname{Mor}_{R-\operatorname{Mod}}(N, E) \xrightarrow{f^{*}} \operatorname{Mor}_{R-\operatorname{Mod}}(M, E) \longrightarrow 0
$$

is an exact sequence of abelian groups
(c) For any injective morphism of $R$-modules $f: M \mapsto N$, and morphism $\phi: M \rightarrow E$, there exists a morphism $\psi: N \rightarrow E$-we do not require uniqueness-for which the diagram

is commutative in $R$-Mod.
(d) Any short exact sequence of $R$-modules of the form

$$
0 \longrightarrow E \longmapsto A \longrightarrow B \longrightarrow 0
$$

is split.
Proof. The equivalence between (a) and (b) comes exactly from the definition of what an injective module is. We first prove the equivalence between (b) and (c), then the equivalence between (c) and (d):

- (b) $\Rightarrow$ (c). If we let $0 \rightarrow M \stackrel{f}{\mapsto} N$ be any exact sequence of $R$-modules, then by (b) we know that $f^{*}: \operatorname{Mor}_{R-\operatorname{Mod}}(N, E) \rightarrow \operatorname{Mor}_{R-\operatorname{Mod}}(M, E)$ is a surjective morphism. Therefore, if we are given any $\phi \in \operatorname{Mor}_{R-\operatorname{Mod}}(M, E)$, one can choose a $\psi \in \operatorname{Mor}_{R-\operatorname{Mod}}(N, E)$ such that $f^{*}(\psi)=\psi f=\phi$.
- (c) $\Rightarrow$ (b). Let $0 \rightarrow M \stackrel{f}{\mapsto} N$ be any exact sequence of $R$-modules, then if we are given any morphism $\phi: M \rightarrow E$, we can find $\psi: N \rightarrow E$ such that $\phi=f \psi$, but $f \psi=f^{*}(\psi)$, therefore we've just shown that the map $f^{*}$ is surjective.
- (c) $\Rightarrow$ (d). Let $0 \rightarrow E \stackrel{f}{\mapsto} A \rightarrow B \rightarrow 0$ be a short exact sequence of $R$-modules. If we consider the identity morphism $\mathrm{id}_{E}$, by item (c) we are able to find a morphism $\psi: A \rightarrow E$ such that $\psi f=\operatorname{id}_{E}$-therefore $f$ is a split monomorphism and the sequence splits.
- (d) $\Rightarrow$ (c). Let $0 \rightarrow L \stackrel{f}{\mapsto} M$ be an exact sequence of $R$-modules and let $\phi: L \rightarrow E$ be a morphism. We may define an $R$-module

$$
X:=\frac{E \oplus M}{\{(\phi(x),-f(x)): x \in L\}}
$$

and—given the natural projection $\pi: E \oplus M \rightarrow X$ and the canonical inclusion morphisms $\iota_{E}: E \hookrightarrow E \oplus M$ and $\iota_{M}: M \hookrightarrow E \oplus M$-we define $\ell_{E}: E \rightarrow X$ and $\ell_{M}: M \rightarrow X$ by $\ell_{E}:=\pi \iota_{E}$ and $\ell_{M}:=\pi \iota_{M}$. By Proposition 8.4.66, we know that $\left(X, \ell_{E}, \ell_{M}\right)$ is the pushout of the pair $(f, \phi)$ :


Let $e \in \operatorname{ker} \ell_{E}$ be any element, then

$$
\ell_{E}(e)=[e, 0] \in\{[\phi(x),-f(x)]: x \in L\} .
$$

Let $x_{e} \in L$ be such that $(e, 0)=\left(\phi\left(x_{e}\right),-f\left(x_{e}\right)\right)$, then $x_{e} \in \operatorname{ker} f$-but since $f$ is injective, it follows that $x_{e}=0$. Therefore $e=\phi(0)=0$.
From this we can consider the following short exact sequence

$$
0 \longrightarrow E \xrightarrow{\ell_{E}} X \longrightarrow \text { coker } \ell_{E} \longrightarrow 0
$$

which by item (d) is split-therefore there exists a retract $r: X \rightarrow E$ of $\ell_{E}$, that is, $r \ell_{E}=\mathrm{id}_{E}$. From this we can create a morphism $\psi: M \rightarrow E$ given by $\psi:=r \ell_{M}-$ we'll show that this map satisfies the condition for item (c). Let $x \in L$ be any element and notice that

$$
\psi f(x)=\left(r \ell_{M}\right) f(x)=r[0, f(x)]=r[\phi(x), 0]=r\left(\ell_{E}(\phi(x))\right)=\left(r \ell_{E}\right) \phi(x)=\phi(x),
$$

therefore $\psi f=\phi$ as needed, which proves that $E$ is injective.

Corollary 10.6.3. Let $M$ be an $R$-module, and $E$ be a submodule of $M$. If $E$ is injective, then it is a direct summand of $M$.

Proof. Indeed, since the short exact sequence

is split, where $\rho$ is a section of $\pi$. Therefore

$$
M=\operatorname{im} \iota \oplus \operatorname{im} \rho=E \oplus \operatorname{im} \rho
$$

Proposition 10.6.4. Any direct summand of an injective $R$-module is an injective $R$-module
Proof. Let $E=X \oplus Y$, and consider a morphism $\phi: L \rightarrow X$, and an injective morphism $f: L \mapsto M$. If $\iota_{X}: X \hookrightarrow E$, we can define a morphism $\phi^{\prime}: L \rightarrow E$ given by $\phi^{\prime}:=\iota_{X} \phi$. Since $E$ is injective, there exists $\psi^{\prime}: M \rightarrow E$ such that $\psi^{\prime} f=\phi^{\prime}$. However, taking the canonical projection $\pi_{X}: E \rightarrow X$ we can define $\psi: M \rightarrow X$ to be given by $\psi:=\pi_{X} \psi^{\prime}$ therefore the diagram

is commutative in $R$-Mod, hence $X$ is injective.
Proposition 10.6.5. Let $\left(E_{j}\right)_{j \in J}$ be a collection of $R$-modules. Then the product $\prod_{j \in J} E_{j}$ is an injective $R$-module if and only if $E_{j}$ is injective for each $j \in J$.

Proof. $(\Rightarrow)$ We know that any permutation $\sigma: J \stackrel{\simeq}{\leftrightharpoons} J$ is such that $\prod_{j \in J} E_{j} \simeq \prod_{j \in J} E_{\sigma(j)}$ in $R$-Mod. Therefore, for any $i \in J$ we have a natural isomorphism

$$
\prod_{j \in J} E_{j} \simeq E_{i} \times\left(\prod_{j \in J \backslash i} E_{j}\right)
$$

which by Proposition 10.6.4 implies in the injectivity of $E_{i}$.
$(\Leftarrow)$ Let $E_{j}$ be injective for each $j \in J$. Let $f: L \leadsto M$ be any injective morphism of $R$-modules and consider a morphism $\phi: L \rightarrow E$. Since each $E_{j}$ is injective, define $\left(\psi_{j}: M \rightarrow E_{j}\right)_{j \in J}$ to be the collection of morphisms such that the diagram

commutes for all $j \in J$-where $\pi_{j}$ is the natural $j$-th projection. From the universal property of products, the collection $\left(\psi_{j}\right)_{j \in J}$ defines a unique morphism $\psi: M \rightarrow E$ such that $\pi_{j} \psi=\psi_{j}$ for each $j \in J$. Note that $\psi f$ and $\phi$ are equal if and only if each of their projections match, but since

$$
\pi_{j}(\psi f)=\left(\pi_{j} \psi\right) f=\psi_{j} f=\pi_{j} \phi
$$

then $\psi f=\phi$ and the diagram

commutes in $R$-Mod-therefore $\prod_{j \in J} E_{j}$ is an injective module.
Remark 10.6.6. Contrary to the behaviour of projective modules, the direct sum of a collection of injective modules need not be injective.

Lemma 10.6.7. Let $R$ be an integral domain and consider its field of fractions Frac $(R)$ as an $R$-module. The following holds:
(a) Let $I \subseteq \operatorname{Frac}(R)$ be an $R$-submodule of the $R$-module $\operatorname{Frac}(R)$, and let $\phi: I \rightarrow \operatorname{Frac}(R)$ be a morphism of $R$-modules. Then there exists $q \in \operatorname{Frac}(R)$ such that $\phi(y)=q y$ for all $y \in I$.
(b) Let $\mathfrak{a}$ be a submodule (ideal) of $R$. Then any morphism of $R$-modules $\phi: \mathfrak{a} \rightarrow$ $\operatorname{Frac}(R)$ can be extended to a morphism $\bar{\phi}: R \rightarrow \operatorname{Frac}(R)$.

Proof. (a) If $I=0$, then $\phi=0$ and any $q \in \operatorname{Frac}(R)$ satisfies the requirement. On the contrary, suppose $I$ is non-zero, and fix any non-zero $b:=w / z \in I$. If $a:=u / v \in$ $\operatorname{Frac}(R)$ is any other element, let $r:=z v$ so that

$$
b r=b(z v)=(b z) v=w v \in R \quad \text { and } \quad a r=a(z v)=a(v z)=(a v) z=u z \in R,
$$

where we used the commutativity of $R$ in order to have $a r \in R$. Therefore one has

$$
\begin{gathered}
\phi((b a) r)=\phi(b(a r))=\phi(b) \cdot(a r), \\
\phi((b a) r)=\phi((a b) r)=\phi(a(b r))=\phi(a) \cdot(b r) .
\end{gathered}
$$

Thus $\phi(a) b r=\phi(b) a r$ —multiplying this equality by $(b r)^{-1} \in \operatorname{Frac}(R)$ we obtain

$$
\phi(a)=\phi(b) \cdot(a r)(b r)^{-1}=\phi(b) \cdot\left(a b^{-1}\right)=\phi(b) \cdot\left(b^{-1} a\right)=\left(\phi(b) \cdot b^{-1}\right) a .
$$

Therefore we may define $q:=\frac{\phi(b)}{b} \in \operatorname{Frac}(R)$, so that $\phi(y)=q y$.
(b) Given any morphism of $R$-modules $\phi: \mathfrak{a} \rightarrow \operatorname{Frac}(R)$, let $q \in \operatorname{Frac}(\underline{R})$ be such that $\phi(y)=q y$ —which exists because of the last items's result. Define $\bar{\phi}: R \rightarrow \operatorname{Frac}(R)$ to be an $R$-module morphism given by $\bar{\phi}(1):=q$-this completely defines $\psi$ since $\bar{\phi}(r)=\bar{\phi}(1 \cdot r)=\bar{\phi}(1) r=q r$. Therefore the diagram


Theorem 10.6.8 (Baer's criterion). Let $E$ be a right- $R$-module (or left). Then $E$ is injective if and only if for every right ideal (or left) $\mathfrak{a}$ of $R$, and morphism of right- $R$-modules (or left) $\phi: \mathfrak{a} \rightarrow E$, there exists an extension $\bar{\phi}: R \rightarrow E$ of $\phi$-that is, the diagram

is commutative in $\operatorname{Mod}_{R}$ ( or $_{R} \operatorname{Mod}$ ).
Proof. $(\Rightarrow)$ If $E$ is injective, then the existence of the extension follows immediately.
$(\Leftarrow)$ Suppose the latter condition is satisfied. Consider morphisms $\phi: L \rightarrow E$, and $f: L \mapsto M$ injective, in $\operatorname{Mod}_{R}$. Define $\mathcal{L}$ to be the collection of all pairs ( $L^{\prime}, \psi^{\prime}: L^{\prime} \rightarrow E$ ) such that $f(L) \subseteq L^{\prime} \subseteq M$, where $L^{\prime}$ is a submodule of $M$, and $\psi^{\prime} f=\phi$. Notice that since $f$ is injective we may define $\psi^{\prime}:=\phi \circ\left(\left.f\right|_{f(L)}\right)^{-1}$ and take $L^{\prime}:=f(L)$, proving that $\mathcal{L}$ is a non-empty set.

We define a partial order $\preccurlyeq$ on $\mathcal{L}$ as follows: $\left(L^{\prime}, \psi^{\prime}\right) \preccurlyeq\left(L^{\prime \prime}, \psi^{\prime \prime}\right)$ if and only if $L^{\prime} \subseteq L^{\prime \prime}$ and $\left.\psi^{\prime \prime}\right|_{L^{\prime}}=\psi^{\prime}$. Define an ascending chain (with respect to $\preccurlyeq$ ) of pairs $\left(L_{j}^{\prime}, \psi_{j}^{\prime}\right)_{j \in J}$ of elements of $\mathcal{L}$, and let $L_{\mathrm{M}}^{\prime}:=\bigcup_{j \in J} L_{j}^{\prime}$. From construction, one has that $f(L) \subseteq L_{\mathrm{M}}^{\prime} \subseteq M$, where $L_{M}^{\prime}$ is a submodule of $M$. Define a morphism $\psi_{M}^{\prime}: L_{M}^{\prime} \rightarrow E$ as follows: if $x \in L_{j}^{\prime}$, define $\psi_{M}^{\prime}(x):=\psi_{j}^{\prime}(x)$ —which is well defined because given $\left(L_{j}^{\prime}, \psi_{j}^{\prime}\right) \preccurlyeq\left(L_{i}^{\prime}, \psi_{i}^{\prime}\right)$, one has $\left.\psi_{i}\right|_{L_{j}}=\psi_{j}$. From its construction the pair $\left(L_{\mathrm{M}^{\prime}}^{\prime}, \psi_{\mathrm{M}}^{\prime}\right)$ is a maximal element of the chain $\left(L_{j}^{\prime}, \psi_{j}^{\prime}\right)$.

Since $\mathcal{L}$ is non-empty and every chain of elements has a maximal element, we can use Zorn's lemma to conclude that $\mathcal{L}$ admits a maximal element $\left(L_{0}, \psi_{0}\right)$. If, for the sake of contradiction, there exists an element $x \in M \backslash L_{0}$, define a right-ideal

$$
\mathfrak{a}:=\left\{r \in R: x r \in L_{0}\right\} \subseteq R,
$$

and a morphism $\lambda: \mathfrak{a} \rightarrow E$ by $r \mapsto \psi_{0}(x r)=\psi_{0}(x) r$. From hypothesis, $\lambda$ admits an extension $\bar{\lambda}: R \rightarrow E$ making the diagram

commute in $\operatorname{Mod}_{R}$. Define a map $\psi_{1}: L_{0}+x R \rightarrow E$ by $\psi_{1}(y+x r):=\psi_{0}(y)+\bar{\lambda}(r)$, which is certainly a morphism of right- $R$-modules. To prove that $\psi_{1}$ is well defined, if $y+x r=y^{\prime}+x r^{\prime} \in L_{0}+x R$ then $y-y^{\prime}=x\left(r^{\prime}-r\right) \in x R$, therefore $y-y^{\prime} \in L_{0}$, while $x\left(r^{\prime}-r\right) \in x R$. This implies in $r^{\prime}-r \in \mathfrak{a}$, therefore both $\psi_{0}\left(y-y^{\prime}\right)$ and $\lambda\left(r^{\prime}-r\right)$. Finally, we see that

$$
\psi_{0}\left(y-y^{\prime}\right)=\psi_{0}\left(x\left(r^{\prime}-r\right)\right)=\lambda\left(r^{\prime}-r\right)=\bar{\lambda}\left(r^{\prime}-r\right)=\bar{\lambda}\left(r^{\prime}\right)-\bar{\lambda}(r)
$$

therefore $\psi_{1}(y+x r)=\psi_{1}\left(y^{\prime}+x r^{\prime}\right)$, thus $\psi_{1}$ is indeed well defined morphism. From construction, we have $L_{0} \subseteq L_{0}+x R$ and $\left.\psi_{1}\right|_{L_{0}}=\psi_{0}$, which implies in

$$
\left(L_{0}, \psi_{1}\right) \preccurlyeq\left(L_{0}+x R, \psi_{1}\right),
$$

contradicting the maximality of $\left(L_{0}, \psi_{1}\right)$. Therefore it must be the case that $L_{0}=M$, which implies in $\psi_{0}: M \rightarrow E$ being such that $\psi_{0} f=\phi$. We conclude that the morphism of right- $R$-modules $\psi_{0}$ makes the diagram

commutative in $\operatorname{Mod}_{R}$, showing that $E$ is an injective module.
Corollary 10.6.9. Let $R$ be an integral domain. Then the following holds:
(a) The field of fractions $\operatorname{Frac}(R)$ is an injective $R$-module
(b) Every $\operatorname{Frac}(R)$-module is an injective $R$-module.
(c) Every vector space is injective.

Proof. (a) Notice that by Lemma 10.6 .7 item (b) we find that $\operatorname{Frac}(R)$ satisfies the lifting property of injective modules described by Baer's criterion, therefore $\operatorname{Frac}(R)$ is an injective $R$-module.
(b) Let $M$ be a $\operatorname{Frac}(R)$-module, and $\mathfrak{a} \subseteq R$ be a non-zero ideal of $R$-if $\mathfrak{a}$ where zero, then the extension property would be satisfies right away. Since $M$ is a module over a field, it's a free module and we can assume that $M \simeq \bigoplus_{j \in J} \operatorname{Frac}(R)$ for some set $J$.

Let $\phi: \mathfrak{a} \rightarrow M$ be any morphism of $R$-modules. If $\pi_{j} \in \operatorname{Mor}_{R-\operatorname{Mod}}(M, \operatorname{Frac}(R))$ denotes the canonical projection of the $j$-th coordinate, we define a collection $\left(f_{j}: \mathfrak{a} \rightarrow \operatorname{Frac}(R)\right)_{j \in J}$ of $R$-module morphisms by $f_{j}:=\pi_{j} \phi$ for each $j \in J$. Notice that since

$$
\phi(a)=\left(f_{j}(a)\right)_{j \in J}=\left(f_{j}(1) a\right)_{j \in J} \in \bigoplus_{j \in J} \operatorname{Frac}(R) \simeq M
$$

it must be the case that $f_{j}(1)$ is non-zero only for finitely many $j \in J$. Therefore $u:=\left(f_{j}(1)\right)_{j \in J}$ is an element of $M$ and we can define completely define a morphism
of $R$-modules $\bar{\phi}: R \rightarrow M$ by mapping $\bar{\phi}(1):=u$. We conclude that such map extends $\phi$, that is

is a commutative diagram in $R$-Mod—which, by Baer's criterion, proves that $M$ is injective.
(c) Given a $k$-vector space $V$, where $k$ is any field, we know that $k=\operatorname{Frac}(k)$. Therefore $V$ is an injective $k$-module by item (b).

Proposition 10.6.10 (Schanuel's dual (see Proposition 10.5.9)). Let $E$ and $E^{\prime}$ be injective $R$-modules. If there exists short exact sequences

$$
\begin{aligned}
& 0 \longrightarrow M \xrightarrow{f} E \xrightarrow{g} L \longrightarrow 0 \\
& 0 \longrightarrow M \xrightarrow{\phi} E^{\prime} \xrightarrow{\psi} L^{\prime} \longrightarrow 0
\end{aligned}
$$

of $R$-modules, then there exists an isomorphism of $R$-modules

$$
E \oplus L^{\prime} \simeq E^{\prime} \oplus L
$$

Proof. From the injectivity of $E^{\prime}$, there exists a morphism of $R$-modules $\varepsilon: E \rightarrow E^{\prime}$ such that the diagram

commutes in $R$-Mod. Define a collection $\left(e_{\ell}\right)_{\ell \in L}$ where $g\left(e_{\ell}\right)=\ell$-which is possible because $g$ is surjective-and define the map $\lambda: L \rightarrow L^{\prime}$ to be given by $\ell \mapsto \psi \varepsilon(\ell)$, then $\lambda$ is certainly a morphism of $R$-modules and $\lambda g=\psi \varepsilon$. We've constructed the following commutative diagram

in $R$-Mod. Define mappings $\alpha: E \rightarrow E^{\prime} \oplus L$ given by $e \mapsto(\varepsilon(e), g(e))$, and $\beta: E^{\prime} \oplus L \rightarrow L^{\prime}$ by $\left(e^{\prime}, \ell\right) \mapsto \lambda(\ell)-\psi\left(e^{\prime}\right)$. From construction, both are $R$-module morphisms. Moreover, notice that $\left(e^{\prime}, \ell\right) \in \operatorname{ker} \beta$ if and only if $\lambda(\ell)=\psi\left(e^{\prime}\right)$, hence if we take $e \in E$ such that $g(e)=\ell$, we find by the commutativity of the right square that

$$
\psi \varepsilon(e)=\lambda g(e)=\lambda(\ell)=\psi\left(e^{\prime}\right)
$$

Therefore $\varepsilon(e)-e^{\prime} \in \operatorname{ker} \psi$ and by exactness there exists $m \in M$ for which $\phi(m)=$ $\varepsilon(e)-e^{\prime}$. From the commutativity of the left square one has

$$
\varepsilon f(m)=\varepsilon(e)-e^{\prime},
$$

thus $\varepsilon(e-f(m))=e^{\prime}$. Notice that

$$
\alpha(e-f(m))=(\varepsilon(e-f(m)), g(e-f(m)))=\left(e^{\prime}, g(e)-g f(m)\right)=\left(e^{\prime}, g(e)\right)=\left(e^{\prime}, \ell\right)
$$

which proves that $\left(e^{\prime}, \ell\right) \in \operatorname{im} \alpha$. Therefore $\operatorname{ker} \beta \subseteq \operatorname{im} \alpha$. For the converse, given any $e \in E$ we have

$$
\beta \alpha(e)=\beta(\varepsilon(e), g(e))=\lambda g(e)-\psi \varepsilon(e)=0
$$

since $\lambda g=\psi \varepsilon$-thus im $\alpha \subseteq \operatorname{ker} \beta$.
We now show that $\alpha$ is injective, while $\beta$ is surjective. If $e \in \operatorname{ker} \alpha$ then by definition $(\varepsilon(e), g(e))=0$. Since $\operatorname{im} f=\operatorname{ker} g$, let $m \in M$ be such that $f(m)=e$. By the commutativity of the left square we know that

$$
\phi(m)=\varepsilon f(m)=\varepsilon(e)=0,
$$

but $\phi$ is injective, thus $m=0$. This proves that $e=f(m)=0$ and therefore $\operatorname{ker} \alpha=0$. For surjectivity, let $\ell^{\prime} \in L^{\prime}$ be any element. Since $\psi$ is surjective, choose $e^{\prime} \in E^{\prime}$ with image $\psi\left(e^{\prime}\right)=\ell^{\prime}$. Taking the pair $\left(-e^{\prime}, 0\right) \in E^{\prime} \oplus L$ we get

$$
\beta\left(-e^{\prime}, 0\right)=\lambda(0)-\psi\left(-e^{\prime}\right)=\psi\left(e^{\prime}\right)=\ell
$$

therefore $\beta$ is surjective.
In the last two paragraphs we've shown that the sequence of $R$-modules

$$
0 \longrightarrow E \xrightarrow{\alpha} E^{\prime} \oplus L \xrightarrow{\beta} L^{\prime} \longrightarrow 0
$$

is short exact. Since $E$ is injective, then the sequence splits, proving the existence of an isomorphism $E^{\prime} \oplus L \simeq E \oplus L^{\prime}$.

## Divisible Modules

Definition 10.6.11 (Divisible module). Let $R$ be a domain, and $D$ be an $R$-module. We say that $D$ is a divisible $R$-module if for all $x \in D$ and $r \in R \backslash\{0\}$, there exists $y \in D$ for which $y r=x$.

Proposition 10.6.12. Let $R$ be an integral domain. The following holds:
(a) The field of fractions $\operatorname{Frac}(R)$ is a divisible $R$-module.
(b) The sum (either direct or not) of divisible $R$-modules is a divisible $R$-module.
(c) The direct product of divisible $R$-modules is a divisible $R$-module.
(d) If $D$ is a divisible $R$-module and $\phi: D \rightarrow M$ is a morphism of $R$-modules, then the image $\phi(D) \subseteq M$ is a divisible $R$-module.
(e) The quotient of a divisible $R$-module is a divisible $R$-module.
(f) Every direct summand of a divisible $R$-module is itself a divisible $R$-module.

Proof. (a) Given any element $x:=u / v \in \operatorname{Frac}(R)$ and a non-zero $r \in R$, one has $y:=u /(v r) \in \operatorname{Frac}(R)$ such that

$$
y r=\frac{u}{v r} \cdot r=\frac{u}{v}=x .
$$

(b) Let $\left(D_{j}\right)_{j \in J}$ be a family of divisible $R$-modules. First we prove that the sum of the family is divisible. Let $\sum_{j \in F} x_{j} \in \sum_{j \in J} D_{j}$ be any element, where $F \subseteq J$ is a finite subset. If $r \in R$ is any non-zero element, since $D_{j}$ is divisible, let $y_{j} \in D_{j}$ be such that $y_{j} r=x_{j}$. Since $F$ is finite, we have $\sum_{j \in F} y_{j} \in \sum_{j \in J} D_{j}$, therefore

$$
\left(\sum_{j \in F} y_{j}\right) r=\sum_{j \in F} y_{j} r=\sum_{j \in F} x_{j} .
$$

This proves that $\sum_{j \in J} D_{j}$ is a divisible $R$-module.
We now consider the direct sum $\bigoplus_{j \in J} D_{j}$ and any element $\left(x_{j}\right)_{j \in J} \in \bigoplus_{j \in J} D_{j}$. If $r \in R$ is non-zero, define a collection $\left(y_{j}\right)_{j \in J}$ as follows: if $x_{j}=0$, let $y_{j}=0 \in D_{j}$, otherwise we use the divisibility of $D_{j}$ and let $y_{j} \in D_{j}$ be an element such that $y_{j} r=$ $x_{j}$. Since finitely many $j \in J$ have a non-zero $x_{j}$, it follows that $\left(y_{j}\right)_{j \in J} \in \bigoplus_{j \in J} D_{j}$. Therefore we have

$$
\left(y_{j}\right)_{j \in J} r=\left(y_{j} r\right)_{j \in J}=\left(x_{j}\right)_{j \in J},
$$

proving that $\bigoplus_{j \in J} D_{j}$ is a divisible $R$-module.
(c) Let $\left(D_{j}\right)_{j \in J}$ be a collection of divisible $R$-modules, and let $\left(x_{j}\right)_{j \in J} \in \prod_{j \in J} D_{j}$ be any element. If $r \in R$ is a non-zero element, we define a collection $\left(y_{j}\right)_{j \in J} \in \prod_{j \in J} D_{j}$ such that $y_{j} r=x_{j}$-which is possible since each $D_{j}$ is divisible. Therefore $\left(y_{j}\right)_{j \in j} r=$ $\left(y_{j} r\right)_{j \in J}=\left(x_{j}\right)_{j \in J}$, proving that the direct product $\prod_{j \in J} D_{j}$ is a divisible $R$-module.
(d) Let $m \in \phi(D)$ be any element and take $d \in \phi^{-1}(m)$. Given any non-zero $r \in R$, since $D$ is divisible, let $y \in D$ be such that $y r=d$. Applying $\phi$ to such element,

$$
m=\phi(d)=\phi(y r)=\phi(y) r .
$$

Since $\phi(y) \in \phi(D)$, this shows that $m$ is divisible by $r$-hence $\phi(D)$ is a divisible $R$-module.
(e) Given a divisible $R$-module $D$ and a submodule $Q \subseteq D$, the natural projection $\pi: D \rightarrow D / Q$ shows that $\pi(D)=D / Q$ is a divisible $R$-module via last item's result.
(f) Let $D$ be a divisible $R$-module and suppose that $D=X \oplus Y$. Then the natural projection $\pi_{X}: D \rightarrow X$ has an image $\pi_{X}(D)=X$-therefore $X$ is divisible by the result of item (d).

Proposition 10.6.13 (Injective module is divisible). Let $R$ be an integral domain. Then every injective $R$-module is a divisible $R$-module.

Proof. Let $E$ be an injective $R$-module, and $x \in E$ be any element. If $r \in R$ is any nonzero element, consider the submodule (ideal) $r R \subseteq R$-which is free with a basis $\{r\}$. By the injectivity of $E$, if $\phi: r R \rightarrow E$ is defined by $\phi(r):=x$, there exists an extension $\psi: R \rightarrow E$ such that $\left.\psi\right|_{r R}=\phi$. Therefore one has

$$
x=\psi(r)=\psi(1 \cdot r)=\psi(1) r,
$$

which proves that $x$ is divisible by $r$, and hence $E$ itself is a divisible $R$-module.
Proposition 10.6.14 (Divisible modules in PIDs). Let $R$ be a principal ideal domain. The following holds:
(a) An $R$-module is injective if and only if it is divisible.
(b) Every quotient of an injective $R$-module is itself injective.

Proof. (a) If $R$ is a principal ideal domain, then in particular it's an integral domain, therefore any injective $R$-module is divisible. We prove the converse. Let $D$ be a divisible $R$-module and take any non-zero submodule (ideal) $\mathfrak{a}$ of $R$, together with a morphism of $R$-modules $\phi: \mathfrak{a} \rightarrow E$. Since $R$ is a PID, assume $\mathfrak{a}=a R$ for some non-zero $a \in R$. By the divisibility of $D$, we know that there exists $d \in D$ such that $\phi(a)=d a$-that is, $\phi(a) \in D$ is divisible by $d$. We define a map $\psi: R \rightarrow D$ by mapping $r \mapsto d r$, which is clearly a morphism of $R$-modules. Notice that if $x:=a r \in \mathfrak{a}$ is any element, then

$$
\psi(x)=\psi(a r)=\psi(a) r=(d a) r=\phi(a) r=\phi(a r)=\phi(x)
$$

that is, $\psi$ extends $\phi$, making the diagram

commutative in $R$-Mod and showing that $D$ is injective.
(b) Let $E$ be an injective module, hence divisible since $R$ is in particular an integral domain. Therefore, if $M \subseteq E$ is any submodule, we know that $E / M$ is a divisible $R$-module by Proposition 10.6.12 item (e). Therefore, by the fact that $R$ is a PID, we use last item's result to obtain that $E / M$ is injective.

Corollary 10.6.15 (Embedding abelian groups into injective ones). Every abelian group can be embedded as a subgroup of an injective abelian group.
Proof. Let $M$ be an abelian group (Z-module) and by Theorem 10.2 .20 we know that there exists a free abelian group $F:=\bigoplus_{j \in J} \mathbf{Z}$ such that $M=F / K$ for some subgroup $K \subseteq F$. Considering the natural inclusion morphism $t: \mathbf{Z} \hookrightarrow \mathbf{Q}$ of $\mathbf{Z}$-modules, we find that $F$ is a submodule of the $\mathbf{Z}$-module $\bigoplus_{j \in J} \mathbf{Q}$. Since $\mathbf{Q}$ is a divisible $\mathbf{Z}$-module, by Proposition 10.6 .12 we find that $E:=\left(\bigoplus_{j \in J} \mathbf{Q}\right) / K$ is again a divisible Z-module. Since $\mathbf{Z}$ is a PID, by Proposition 10.6 .14 we find that $E$ is injective. Therefore the proposition follows by the natural inclusion $M=F / K \hookrightarrow E$.

Theorem 10.6.16 (Embedding modules into injective ones). Every right- $R$-module (or left) can be embedded as a submodule of some injective right-R-module (or left).

Proof. Let $M$ be any right- $R$-module. By Corollary 10.6.15, let $D$ be an injective abelian group containing the abelian group $M$. Considering the isomorphism $M \simeq$ $\operatorname{Mor}_{\text {Mod }_{R}}(R, M)$ given by $m \mapsto f_{m}$-where $f_{m}(r):=m r$-we have an injective right- $R$ module morphism $M \mapsto \operatorname{Mor}_{\text {Modz }^{\prime}}(R, D)$ given by the composition:

$$
M \xrightarrow{\longrightarrow} \operatorname{Mor}_{\operatorname{Mod}_{R}}(R, M) \longleftrightarrow \operatorname{Mor}_{\operatorname{Mod}_{Z}}(R, M) \longleftrightarrow \operatorname{Mor}_{\operatorname{Mod}_{Z}}(R, D)
$$

This proves that $M$ can be embedded as a subgroup of the right- $R$ - module $\operatorname{Mor}_{\text {Mod }}(R, D)$.

Proposition 10.6.17. Let $D$ be a divisible abelian group and $R$ be any ring. Then the abelian group $\operatorname{Mor}_{\text {Modz }_{\mathrm{z}}}(R, D)$ is an injective right- $R$-module $\left(\operatorname{also} \operatorname{Mor}_{\mathrm{Z}} \operatorname{Mod}(R, D)\right.$ is an injective left- $R$-module), with a product

$$
(f \cdot r)(x):=f(r x)
$$

for every $f \in \operatorname{Mor}_{\text {Modz }_{z}}(R, D)$, and $r, x \in R$.
Proof. We take $D$ to be a right-Z-module-which is isomorphic to the left module since $\mathbf{Z}$ is commutative. Any ring $R$ has the structure of $(R, \mathbf{Z})$-bimodule, thus the product is well defined and produces a structure of right- $R$-module in the abelian group $\operatorname{Mor}_{\text {Modz }_{z}}(R, D)$.

We now prove that $\operatorname{Mor}_{\text {Modz }}(R, D)$ is injective using Baer's criterion. Since $\mathbf{Z}$ is a PID, the divisibility of $D$ implies that it is an injective Z-module. Let $\mathfrak{a} \subseteq R$ be any right-ideal of $R$ and $\phi: \mathfrak{a} \rightarrow \operatorname{Mor}_{\text {Mod }_{Z}}(R, D)$ be a morphism of $R$-modules. Define a map $\phi^{\prime}: \mathfrak{a} \rightarrow D$ by $a \mapsto \phi(a)(1)$, which is a morphism of right-Z-modules. By the injectivity of $D$, there exists a morphism of right-Z-modules $\psi^{\prime}: R \rightarrow D$ such that the diagram

commutes in Modz. Define a morphism of right- $R$-modules $\psi: R \rightarrow \operatorname{Mor}_{\operatorname{Mod}_{\mathbf{Z}}}(R, D)$ by assigning $\psi(1):=\psi^{\prime}$, which completely determines $\psi$ since

$$
\psi(s)(r)=(\psi(1) s)(r)=\left(\psi^{\prime} s\right)(r)=\psi^{\prime}(s r)
$$

for any $s, r \in R$. Moreover, if $a \in \mathfrak{a}$ is any element, then for any $r \in R$ one has

$$
\psi(a)(r)=\psi^{\prime}(a r)=\phi^{\prime}(a r)=\phi(a r)(1)=(\phi(a) r)(1)=\phi(a)(r)
$$

since $a r \in \mathfrak{a}$. Therefore $\left.\psi\right|_{\mathfrak{a}}=\phi$, making the diagram

commute in $\operatorname{Mod}_{R}$ —which shows that $\operatorname{Mor}_{\text {Modz }}(R, D)$ is an injective right- $R$-module. $\quad$ 而

Lemma 10.6.18. Let $E$ be an injective $R$-module, and $0 \rightarrow A \stackrel{\alpha}{\rightarrow} B \xrightarrow{\beta} C \rightarrow 0$ be a short exact sequence of $R$-modules. If there exists a morphism $\gamma: A \rightarrow E$, then one can complete the given exact sequence to a commutative diagram in $R$-Mod:

whose rows are short exact sequences.
Proof. Let $P$ be the pushout of the pair $(\gamma, \alpha)$, that is:

$$
P:=\frac{E \oplus B}{\{(\gamma(a),-\alpha(a)) \in E \oplus B: a \in A\}}
$$

together with natural inclusion maps $\alpha^{\prime}: E \rightarrow P$ given by $e \mapsto[e, 0]$ and $\gamma^{\prime}: B \rightarrow P$ sending $b \mapsto[0, b]$. Define $\beta^{\prime}: P \rightarrow C$ to be the mapping $[e, b] \mapsto \beta(b)$. To see that $\beta^{\prime}$ is well defined, let $[e, b]=\left[e^{\prime}, b^{\prime}\right]$ and notice this means that there exists a common $a \in A$ such that $\alpha(a)=b-b^{\prime}$ and $\gamma(a)=e-e^{\prime}$. Since the top row is exact, then

$$
0=\beta \alpha(a)=\beta\left(b-b^{\prime}\right)=\beta(b)-\beta\left(b^{\prime}\right)
$$

that is, $\beta^{\prime}[e, b]=\beta^{\prime}\left[e^{\prime}, b^{\prime}\right]$. Also it is clear that $\beta^{\prime}$ is a morphism of $R$-modules and that $\beta^{\prime} \gamma^{\prime}=\beta$.

Given any $e \in \operatorname{ker} \alpha^{\prime}$ one has $\alpha^{\prime}(e)=[e, 0]=[0,0]$, that is, there exists $a \in A$ for which $(\gamma(a),-\alpha(a))=(e, 0)$ but since $\alpha$ is injective, then $a=0$-proving that ker $\alpha^{\prime}=0$ and that $\alpha^{\prime}$ is injective. For the surjectivity of $\beta^{\prime}$, we use the fact that $\beta$ is surjective: given any $c \in C$, there exists $b \in B$ for which $\beta(b)=c$-therefore $\beta^{\prime} \gamma^{\prime}(b)=\beta^{\prime}[0, b]=\beta(b)=c$, showing that $\beta^{\prime}$ is surjective.

We now show the exactness of the bottom row. Since $\beta^{\prime} \alpha^{\prime}(e)=\beta^{\prime}[e, 0]=\beta(0)=0$, then $\operatorname{im} \alpha^{\prime} \subseteq \operatorname{ker} \beta^{\prime}$. Given any $[e, b] \in \operatorname{ker} \beta^{\prime}$ by definition we have $\beta(b)=0$, therefore by exactness of the top row there must exist $a \in A$ such that $\alpha(a)=b$. Therefore

$$
[e, b]=[e, \alpha(a)]=[e, \alpha(a)]+[\gamma(a),-\alpha(a)]=[e+\gamma(a), 0]
$$

since $[\gamma(a),-\alpha(a)]=[0,0]$. From this we obtain that $\alpha^{\prime}(e+\gamma(a))=[e, b]$ and hence $\operatorname{ker} \beta^{\prime} \subseteq \operatorname{im} \alpha^{\prime}$.

Proposition 10.6.19. An $R$-module $E$ is injective if and only if every short exact sequence of $R$-modules

$$
0 \longrightarrow E \xrightarrow{f} B \xrightarrow{g} C \longrightarrow 0
$$

ending with a cyclic module $C$ is split.
Proof. The forward implication is immediate since any short exact sequence of right- $R$ modules starting with an injective module is split. For the converse, assume that $E$ has the described property, and let $\mathfrak{b} \subseteq R$ be any right-ideal together with a morphism of
right- $R$-modules $\phi: \mathfrak{b} \rightarrow E$. Considering the short exact sequence $0 \rightarrow \mathfrak{b} \hookrightarrow R \rightarrow R / \mathfrak{b}$, using Lemma 10.6.18 there exists a commutative diagram


Since $R / b$ is a cyclic right- $R$-module, by hypothesis the bottom sequence splits, proving the existence of a retract $q: P \rightarrow E$ such that $q \alpha=\operatorname{id}_{E}$. Therefore the composition $q \phi^{\prime}: R \rightarrow E$ is a morphism of right- $R$-modules such that $(q \phi) \iota=\phi$, making the diagram

commute in $\operatorname{Mod}_{R}$. Thus by Baer's criterion $E$ is an injective module.
Example 10.6.20. The $\mathbf{Z}$-module $\mathbf{Q} / \mathbf{Z}$ is injective.
Proof. Since $\mathbf{Z}$ is a principal integral domain and $\mathbf{Q} / \mathbf{Z}$ is divisible, then $\mathbf{Q} / \mathbf{Z}$ is also injective.

Theorem 10.6.21. Let $R$ be a ring. There exists enough injective modules in the category $R$-Mod-that is, every $R$-module is a submodule of an injective $R$-module.
Write (see stacks project).

### 10.7 Injective Cogenerator Modules

Definition 10.7.1 (Injective cogenerator). An injective $R$-module $Q$ is said to be cogenerator if for all non-zero $R$-modules $M$ we have $\operatorname{Mor}_{R-\operatorname{Mod}}(M, Q) \neq 0$.

Lemma 10.7.2. The $\mathbf{Z}$-module $\mathbf{Q} / \mathbf{Z}$ is an injective cogenerator of $\mathbf{Z}$-Mod.
Proof. Let $M$ be a non-zero Z-module, and let $m \in M \backslash 0$. We have two cases:

- If $m$ has infinite order, define a map $u: \mathbf{Z} m \rightarrow \mathbf{Q} / \mathbf{Z}$ given by $u(x):=1 / 2+\mathbf{Z}$.
- If $m$ has a finite order $n$, define a morphism $u: \mathbf{Z} m \rightarrow \mathbf{Q} / \mathbf{Z}$ to be given by $u(x):=1 / n+\mathbf{Z}$.

Since $\mathbf{Q} / \mathbf{Z}$ is an injective $\mathbf{Z}$-module, the non-zero morphism $u$ can be extended to a morphism $v: M \rightarrow \mathbf{Q} / \mathbf{Z}$, which is also non-zero. Therefore $\operatorname{Mor}_{\mathbf{Z}-\operatorname{Mod}}(M, \mathbf{Q} / \mathbf{Z}) \neq 0$. $\quad$,

Proposition 10.7.3. Let $Q$ be an injective cogenerator $R$-module. For every $R$-module $M$ there exists a set $\Lambda$ and a monomorphism $M \mapsto Q^{|\Lambda|}$.

In other words, if there exists an injective cogenerator object in $R$-Mod, then any module of $R$-Mod can be made a submodule of an injective one.

Proof. Let $\Lambda:=\operatorname{Mor}_{R-\operatorname{Mod}}(M, Q)$, which contains non-zero morphisms since $Q$ is assumed to be an injective cogenerator. Define a map $f: M \rightarrow Q^{|\Lambda|}$ to be given by $x \mapsto(u(x))_{u \in \Lambda}$, which is an $R$-linear morphism. Let $m \in M$ be any non-zero element and take a non-zero morphism $v: R m \rightarrow Q$. By the injectivity of $Q$ we find an extension $\bar{v}: M \rightarrow Q$ of $v$. In particular, since $v$ is non-zero, then $\bar{v}(m) \neq 0$ and therefore $f(m) \neq 0$-showing that $f$ has null-kernel and thus is a monomorphism.

Proposition 10.7.4 (Every category $A$-Mod has an injective cogenerator). Let $A$ be a $k$-algebra, where $k$ is a commutative ring. The right $A$-module $\operatorname{Mor}_{M_{M o d}}\left({ }_{A} A_{\mathbf{Z}}, \mathbf{Q} / \mathbf{Z}\right)$ is an injective cogenerator in $\operatorname{Mod}_{A}$.

Proposition 10.7.5 (Injective resolution). Let $k$ be a commutative ring and $A$ be a $k$-algebra. For any $A$-module $M$ there exists an exact sequence

$$
0 \longrightarrow M \xrightarrow{f_{0}} E_{0} \xrightarrow{f_{1}} E_{1} \longrightarrow \cdots \longrightarrow E_{n} \xrightarrow{f_{n+1}} E_{n+1} \longrightarrow \cdots
$$

where each $A$-module $E_{j}$ is injective. Such sequence is called an injective resolution of $M$.

Definition 10.7.6 (Injective presentation). An injective presentation of an $A$-module $M$ is an injective resolution composed of only two injective modules:

$$
0 \longrightarrow M \succ E_{0} \longrightarrow E_{1}
$$

Lemma 10.7.7. Let $Q$ be an injective cogenerator of $A$-Mod. A sequence of $A$-modules

$$
L \xrightarrow{f} M \xrightarrow{g} N
$$

is exact if and only if the induced sequence of $k$-modules

$$
\operatorname{Mor}_{A-\operatorname{Mod}}(N, Q) \xrightarrow{g^{*}} \operatorname{Mor}_{A-\operatorname{Mod}}(M, Q) \xrightarrow{f^{*}} \operatorname{Mor}_{A-\operatorname{Mod}}(L, Q)
$$

is exact.
Proof. • $(\Rightarrow)$ Since $Q$ is injective, the contravariant functor $\operatorname{Mor}_{A-M o d}(-, Q)$ is exact and therefore if $L \rightarrow M \rightarrow N$ is exact then the induced sequence is also exact.

- $(\Leftarrow)$ Assume that the induced sequence is exact. We must show that $g f=0$. To that end suppose, for the sake of contradiction, that there exists $\ell \in L$ such that $g f \ell \neq 0 \in N$. Since $Q$ is a cogenerator, there exists a non-zero morphism
$\langle g f \ell\rangle \rightarrow Q$ and, using the fact that $Q$ is injective, there exists an extension $u: N \rightarrow Q$ of such non-zero morphism:

which certainly satisfies $u g f \ell \neq 0$. This shows that $f^{*} g^{*}=(g f)^{*}$ is non-zero, which is a contradiction since we are assuming that the induced sequence is exact. This shows that $g f=0$ and thus $\operatorname{im} f \subseteq \operatorname{ker} g$.
To prove the other side of the inclusion, let $m \in \operatorname{ker} g$ be any element and suppose, for the sake of contradiction, that $m \notin \operatorname{im} f$. Define $p: M \rightarrow$ coker $f$ to be the canonical projection, which therefore satisfies $p m \neq 0$. Since $Q$ is an injective cogenerator, there exists a non-zero morphism $\langle m+\operatorname{im} f\rangle \rightarrow Q$ and since $Q$ is injective, there exists an extension $v:$ coker $f \rightarrow Q$ such that the following diagram commutes:

thus in particular $v p m \neq 0$. From definition, $p f=0$ and therefore $f^{*}(v p)=$ $v p f=0$, hence $v p \in \operatorname{ker} f^{*}=\operatorname{im} g^{*}$. Thus there exists $w: N \rightarrow Q$ such that $g^{*} w=w g=v p$, forming the following commutative diagram


However, since we are assuming that $m \in \operatorname{ker} g$ then

$$
0=g m=w g m=v p m \neq 0,
$$

which is a contradiction, therefore it must be the case that $m \in \operatorname{im} f$. This shows us that $\operatorname{ker} g \subseteq \operatorname{im} f$, therefore the sequence is exact.

### 10.8 Essential Extensions and Injective Envelopes

Definition 10.8.1 (Essential extension). Let $M$ be an $A$-module. An $A$-module $E$ is said to be an essential extension of $M$ if the following are satisfied:
(a) $M$ is a submodule of $E$.
(b) Any non-zero submodule $N \subseteq E$ has a non-trivial intersection with $M$-that is, $N \cap M \neq 0$.

We shall also say that $M$ is an essential submodule of $E$.
Example 10.8.2. The $\mathbf{Z}$-module $\mathbf{Q}$ is an essential extension of $\mathbf{Z}$.
Definition 10.8.3 (Essential monomorphism). A monomorphism $f: M \mapsto E$ in $A$-Mod is said to be essential if for any morphism of $A$-modules $g: E \rightarrow F$ such that the composition $g f$ is a monomorphism then $g$ is a monomorphism.

Lemma 10.8.4. Let $f: L \mapsto M$ and $g: M \mapsto L$ be monomorphisms in $A$-Mod. The following properties hold:
(a) If both $f$ and $g$ are essential, then the composition $g f$ is an essential monomorphism.
(b) If the composition $g f$ is an essential monomorphism, then $g$ is an essential monomorphism.

Proof. (a) Let $h: N \rightarrow S$ be a morphism of $A$-modules such that $h g f: L \rightarrow S$ is a monomorphism then $(h g) f$ is a monomorphism and since $f$ is essential then $h g$ is a monomorphism. Using that $g$ is essential we conclude that $h$ is a monomorphism.
(b) Assume that $g f: L \rightarrow N$ is an essential monomorphism. Therefore if $h: N \rightarrow S$ in $A$-Mod is such that $h g$ is a monomorphism then in particular $(h g) f=h(g f)$ is a monomorphism. Since $g f$ is assumed to be essential then $h$ must be a monomorphism. This shows that $g$ is essential.

Proposition 10.8.5. A monomorphism $f: M \mapsto E$ is essential if and only if $E$ is an essential extension of the image $f M$.

Proof. - $(\Rightarrow)$ Assume that $f$ is essential and let $N \subseteq E$ be a submodule such that $N \cap f M=0$. If $p: E \rightarrow E / N$ is the canonical projection, then if $x \in \operatorname{ker}(p f)$ it follows that $f x \in \operatorname{ker} p=N$-therefore $f x \in N \cap f M$ and hence $f x=0$, thus $x \in \operatorname{ker} f=0$. This shows that $p f: M \mapsto E / N$ is a monomorphism, therefore it must be the case that $N=\operatorname{ker}(p f)=0$.

- $(\Leftarrow)$ Assume that $E$ is an essential extension of $f M$ and let $g: E \rightarrow L$ in $A$-Mod be such that $g f: M \mapsto L$ is a monomorphism. Suppose, for the sake of contradiction, that $g$ isn't a monomorphism-that is, $\operatorname{ker} g \neq 0$. Since $\operatorname{ker} g$ is not trivial then $\operatorname{ker} g \cap f M \neq 0$ and there must exist $x \in M$ with $f x \neq 0$ such that $f x \in \operatorname{ker} g$. This contradicts the fact that $g f$ is a monomorphism, since $\operatorname{ker}(g f)=0$. Hence $g$ must be such that $\operatorname{ker} g=0$, thus a monomorphism.

Lemma 10.8.6 (Injectives \& essential extensions). A module is injective if and only if it does not admit any proper essential extensions.

Proof. - $(\Rightarrow)$ Suppose that $M$ is injective and let $E$ be an $A$-module with $M$ as one of its proper submodules. Since $M$ is injective, then the following sequence is split:

$$
0 \longrightarrow M \longleftrightarrow E \longrightarrow E / M \longrightarrow 0
$$

Hence $E \simeq M \oplus E / M$ and $M \cap E / M=0$ and thus $E$ isn't an essential extension of M.

- $(\Leftarrow)$ Let $M$ be an $A$-module that does not admit proper essential extensions. Using Theorem 10.6.21, let $E$ be an injective $A$-module containing $M$. Via Zorn's lemma, the collection of all submodules of $E$ having a trivial intersection with $M$ admits a maximal element $N \subseteq E$. Let $\iota: M \hookrightarrow E$ be the canonical inclusion, and $\pi: E \rightarrow E / N$ be the canonical projection. Since $\operatorname{ker}(\pi \iota)=M \cap N=0$, then $\pi \iota: M \mapsto E / N$ is a monomorphism. By the maximality of $N$ we find that $E / N$ is an essential extension of $M$ : indeed, if $L / N \subseteq E / N$ is a non-zero submodule then $N \subsetneq L$ is a proper submodule and hence $L \cap M \neq 0$ by hypothesistherefore $(L / N) \cap M \neq 0$. By hypothesis $M$ does not admit a proper essential extension, therefore $E / N=M$. This shows that $\pi \iota=\operatorname{id}_{M}$ and therefore $\iota$ is a split monomorphism, proving that $M$ is a direct summand of $E$, and hence $M$ is injective.

Definition 10.8.7 (Maximal essential extension). An essential extension $E$ of $M$ is said to be maximal if, given an essential extension $E^{\prime}$ of $M$ such that $E \subseteq E^{\prime}$ then $E^{\prime}=E$.

Theorem 10.8.8. Let $E$ be an $A$-module, and $M$ be a submodule of $E$. The following conditions are equivalent:
(a) $E$ is a maximal essential extension of $M$.
(b) $E$ is an essential extension of $M$ and is injective.
(c) $E$ is injective and, if $I$ is an injective module such that $M \subseteq I \subseteq E$, then $I=E$.

Furthermore, every $A$-module admits a maximal essential extension.
Proof. - $(\mathrm{a} \Rightarrow \mathrm{b})$ Since $E$ is a maximal essential extension, in particular $E$ does not admit a proper essential extension-therefore by Lemma 10.8.6 it follows that $E$ is injective.

- $(\mathrm{b} \Rightarrow \mathrm{c})$ Suppose, for the sake of contradiction, that there exists an injective $I$ such that $M \subseteq I \subsetneq E$. From injectivity, $I$ is a direct summand of $E$ and since by hypothesis we have $I \neq E$, it follows that there exists a non-zero $I^{\prime} \subseteq E$ such that $E=I \oplus I^{\prime}$. Since $M \subseteq I$ and $I \cap I^{\prime}=0$ then $M \cap I^{\prime}=0$, which is a contradiction from the fact that $E$ is an essential extension.
- $(c \Rightarrow a)$ From Theorem 10.6 .21 there exists an injective $A$-module $J$ such that $M$ is one of its submodules. Define $\mathcal{E}$ to be the collection of all essential extensions of $M$ contained in $J$. Since $M$ itself is an element of $\mathcal{E}$, then $\mathcal{E}$ is a non-empty set which can made a poset via inclusion. It follows from Zorn's lemma that $\mathcal{E}$
admits a maximal element $E$. We now prove that $E$ is a maximal extension of $M$. Suppose, on the contrary, that there exists an essential extension $E^{\prime}$ of $M$ such that $E \subseteq E^{\prime}$. By injectivity of $E$, there exists an extension $t: E^{\prime} \hookrightarrow J$ of the inclusion $i: E \hookrightarrow J$ so that the following diagram commutes:


Therefore $\iota$ is a monomorphism, which shows that $E^{\prime} \subseteq J$ hence $E^{\prime} \in \mathcal{E}$ and by maximality of $E$ it follows that $E^{\prime}=E$. This shows that $E$ is a maximal essential extension of $M$.

### 10.9 Tensor Products

Definition 10.9.1 (Biadditive and bilinear maps). Let $R$ be a ring and consider a right $R$-module $A$, and a left $R$-module $B$, and an abelian group $G$. A map $f: A \times B \rightarrow G$ is said to be $R$-biadditive if, for any pairs $(a, b),\left(a^{\prime}, b^{\prime}\right) \in A \times B$ and scalar $r \in R$ we have the following conditions satisfied:
(a) $f\left(a+a^{\prime}, b\right)=f(a, b)+f\left(a^{\prime}, b\right)$.
(b) $f\left(a, b+b^{\prime}\right)=f(a, b)+f\left(a, b^{\prime}\right)$.
(c) $f(a r, b)=f(a, r b)$.

Further, if $R$ is commutative, we say that $f$ is $R$-bilinear if it also satisfies:
(d) $f(a r, b)=f(a, r b)=r f(a, b)$.

Now let $A$ be a $k$-algebra, where $k$ is a commutative ring, and consider modules $L_{A}$ and ${ }_{A} M$. Then we define a morphism of $k$-modules $g: L \times M \rightarrow X$ to be $A$-bilinear if

- $g\left(\ell \alpha+\ell^{\prime} \beta, m\right)=g(\ell, m) \alpha+g\left(\ell^{\prime}, m\right) \beta$, for any $\alpha, \beta \in k$.
- $g\left(\ell, m \alpha+m^{\prime} \beta\right)=g(\ell, m) \alpha+g\left(\ell, m^{\prime}\right) \beta$, for any $\alpha, \beta \in k$.
- $g(\ell a, m)=g(\ell, a m)$.

Definition 10.9.2 (Tensor product). Let $R$ be a ring and consider right and left $R$ modules $A_{R}$ and ${ }_{R} B$. The tensor product of $A$ and $B$ is an abelian group $A \otimes_{R} B$ together with an $R$-biadditive map $t: A \times B \rightarrow A \otimes_{R} B$ such that, for every abelian group $G$ and any $R$-biadditive map $f: A \times B \rightarrow G$, there exists a unique abelian group morphism $\phi: A \otimes_{R} B \rightarrow G$ such that the following triangle commutes:


Lemma 10.9.3 (Uniqueness up to isomorphism). Let $X$ and $Y$ be tensor products for $R$-modules $A_{R}$ and ${ }_{R} B$, then there exists a natural isomorphism $X \simeq Y$.

Proof. Let $t_{X}: A \times B \rightarrow X$ and $t_{Y}: A \times B \rightarrow Y$ be their respectively associated $R$-biadditive map. Consider the universal property of the tensor product applied to both $\left(X, t_{X}\right)$ and $\left(Y, t_{Y}\right)$ : we obtain unique morphisms of groups $\phi_{1}: X \rightarrow Y$ and $\phi_{2}: Y \rightarrow X$ such that the following diagram commutes


From the left diagram we see that $\mathrm{id}_{X}: X \rightarrow X$ also makes the diagram commute and on the right diagram id $Y: Y \rightarrow Y$ is also a possible morphism satisfying commutativity. Since the morphisms $\phi_{1}$ and $\phi_{2}$ are uniquely chosen, then the compositions $\phi_{1} \phi_{2}$ and $\phi_{2} \phi_{1}$ are also unique. Therefore it follows that $\phi_{1} \phi_{2}=\mathrm{id}_{Y}$ and $\phi_{2} \phi_{1}=\mathrm{id}_{X}$-hence $X \simeq Y$.

Lemma 10.9.4 (Existence). Given a ring $R$ together with right and left $R$-modules $A_{R}$ and ${ }_{R} B$, their tensor product $A \otimes_{R} B$ exists.

Proof. Let $F$ be the free abelian group with basis $A \times B$, and consider the subgroup $S$ of $S$ generated by elements of the following form, for any $(a, b),\left(a^{\prime}, b^{\prime}\right) \in A \times B$ and $r \in R$ :

- $\left(a, b+b^{\prime}\right)-(a, b)-\left(a, b^{\prime}\right)$,
- $\left(a+a^{\prime}, b\right)-(a, b)-\left(a^{\prime}, b\right)$,
- $(a r, b)-(a, r b)$.

We'll show that the quotient $F / S$ is the wanted tensor product $A \otimes_{R} B$. Define the map $t: A \otimes B \rightarrow F / S$ naturally by mapping $(a, b) \mapsto(a, b)+S$. From the definition of $S$ we obtain freely that $t$ is $R$-biadditive.

Let $G$ be any abelian group and $f: A \times B \rightarrow G$ any $R$-biadditive map. Then from the free group theorem there exists a unique morphism of groups $\phi: F \rightarrow G$ determined by mapping $(a, b) \mapsto f(a, b)$. Moreover, $S \subseteq \operatorname{ker} \phi$ from the fact that we imposed $R$ biadditiveness in $f$. Therefore, by the quotient theorem $\phi$ induces a unique morphism of groups $\psi: F / S \rightarrow G$ satisfying $\psi t=f$ :


This shows that $F / S$ satisfies the universal property of the tensor product and thus $A \otimes_{R} B \simeq F / S$.

Proposition 10.9.5. Consider a morphism of left $R$-modules $\lambda: A_{R} \rightarrow B_{R}$ and a morphism of right $R$-modules $\rho:_{R} X \rightarrow{ }_{R} Y$. The maps $\lambda$ and $\rho$ induce a unique morphism of abelian groups

$$
\lambda \otimes \rho: A \otimes_{R} X \longrightarrow B \otimes_{R} Y
$$

This morphism is given by the mapping $a \otimes x \mapsto \lambda(a) \otimes \rho(x)$.
Proof. Let $\phi: A \times X \rightarrow B \otimes_{R} Y$ be the map given by $\phi(a, b):=a \otimes b$. Then for any $r \in R$ one has

$$
\phi(a r, b)=\lambda(a r) \otimes \rho(b)=\lambda(a) r \otimes \rho(b)=\lambda(a) \otimes r \rho(b)=\lambda(a) \otimes \rho(r b)=\phi(a, r b) .
$$

Thus $\phi$ is $R$-biadditive. The morphism uniquely induced morphism $A \otimes_{R} X \rightarrow B \otimes_{R} Y$ is exactly $\lambda \otimes \rho$ as described above.

Corollary 10.9.6. Consider right $R$-module morphisms $A \xrightarrow{\rho_{1}} B \xrightarrow{\rho_{2}} C$ and morphisms of left $R$-modules $X \xrightarrow{\lambda_{1}} Y \xrightarrow{\lambda_{2}} Z$. Then we have a composition

$$
\left(\rho_{2} \otimes \lambda_{2}\right) \circ\left(\rho_{1} \otimes \lambda_{1}\right)=\rho_{2} \rho_{1} \otimes \lambda_{2} \lambda_{1}
$$

Lemma 10.9.7 (Additive tensoring functors). Let $A$ be a right $R$-module and $B$ be a left $R$-module.
(a) There exists an additive functor $F_{A}:{ }_{R} \operatorname{Mod} \rightarrow \mathrm{Ab}$ given by $F_{A}(B)=A \otimes_{R} B$ and $F_{A}(\beta)=\mathrm{id}_{A} \otimes \beta$. We shall denote this functor by $A \otimes_{R}-$.
(b) There exists an additive functor $G_{B}: \operatorname{Mod}_{R} \rightarrow A b$ mapping $G_{B}(A)=A \otimes_{R} B$ and $G_{B}(\alpha)=\alpha \otimes \operatorname{id}_{B}$. This functor will be denoted by $-\otimes_{R} B$.

Corollary 10.9.8 (Induced isomorphisms). Let $f \in \operatorname{Mor}_{\operatorname{Mod}_{R}}(M, N)$ and $g \in \operatorname{Mor}_{R} \operatorname{Mod}(L, K)$ be isomorphisms. The induced morphism $f \otimes g: M \otimes_{R} L \rightarrow N \otimes_{R} K$ is an isomorphism of abelian groups.

Proof. The functor $-\otimes_{R} L$ takes $f$ to the isomorphism $f \otimes \operatorname{id}_{L}$ and $M \otimes_{R}$ - takes $g$ to the isomorphism $\operatorname{id}_{M} \otimes g$. The composition $\left(f \otimes \operatorname{id}_{L}\right)\left(\operatorname{id}_{M} \otimes g\right)=f \otimes g$ will thus be an isomorphism.

Proposition 10.9.9 (Extending scalars). Let $R$ be a ring and consider a subring $S \subseteq R$.
(a) Let $A \in(R, S)$-Mod, then for any $B \in{ }_{s}$ Mod the tensor product $A \otimes_{S} B$ inherits a natural structure of a left $R$-module: for any $r \in R$ and $a \otimes b \in A \otimes_{S} B$ we have $r(a \otimes b)=(r a) \otimes b$.
(b) Let $B \in(S, R)$-Mod, then for any $A \in \operatorname{Mod}_{S}$ the tensor product $A \otimes_{S} B$ inherits a natural structure of a right $R$-module: for any $r \in R$ and $a \otimes b \in A \otimes_{S} B$ we have $(a \otimes b) r=a \otimes(b r)$.

Proof. We shall prove only the first proposition since the other follows analogously. Given any $r \in R$, consider the left multiplication morphism ${ }_{r}$ mul: $A \rightarrow A$ and the morphism of groups ${ }_{r} \mathrm{mul}_{\otimes} \mathrm{id}_{B}$ under the functor $-\otimes_{S} B$. Notice that this map associates $a \otimes b \mapsto_{r} \operatorname{mul}(a) \otimes b=(r a) \otimes b$. It is trivial to verify that ${ }_{r}$ mul induces a multiplicative left $R$-module structure on the abelian group $A \otimes_{S} B$-transforming it into a left $R$-module.

Example 10.9.10. If $R$ is ring and $S$ is any subring of $R$, then $R$ has a structure of $(R, S)$ bimodule and therefore for any given left $S$-module $M$ we can obtain a left $R$-module $R \otimes_{S} M$. That is, the procedure shown in Proposition 10.9.9 allows one to extend the scalars associated to a module.

## Part III

## Combinatorics

## Chapter 11

## A Short Toy-Manual of Graphs

### 11.1 Basic Principles of Combinatorics

Notation 11.1.1 ( $k$-subsets). Given a set $A$, we define a $k$-subset-where $0 \leqslant k \leqslant|A|$ to be the a subset of $A$ with exactly $k$ elements. We denote the collection of all $k$-subsets of $A$ as $[A]^{k}$.

Notation 11.1.2 (Range). For any $n \in \mathbf{N}$, we denote by $[n$ ] the set of natural numbers $\{1, \ldots, n\}$.

Definition 11.1.3 (Partition). Given a set $V$, a partition is a collection $V_{1}, \ldots, V_{n}$ such that for each $v \in V$ there exists a unique index $1 \leqslant i \leqslant n$ such that $v \in V_{i}$.

Definition 11.1.4 (Colouring). Given a set $V$ we define an $r$-colouring of $V$, for $r \in \mathbf{N}$, to be a function $V \rightarrow[r]$.

Axiom 11.1.5 (Pigeon hole principle). Given a collection $V$ of $n$ elements and an $r$ colouring $c: V \rightarrow[r]$, with $r<n$, then there exists at least one colour $j \in[r]$ such that $c(v)=j=c\left(v^{\prime}\right)$ for distinct elements $v, v^{\prime} \in V$.

Lemma 11.1.6 (Stifel relation). Given numbers $n, k \geqslant 1$ with $k \leqslant n$, we have

$$
\binom{n+1}{k+1}=\binom{n}{k}+\binom{n}{k+1}
$$

Proof. The left side of the equation counts the number of sets with $k+1$ elements out of $[n+1]$. For the interpretation of the right side, let $x \in[n+1]$ be any fixed element, then consider:

- The number of sets composed of $k+1$ elements of $[n+1] \backslash x \simeq[n]$ is given by $\binom{n}{k+1}$.
- The number of sets composed of $k+1$ elements of $[n+1]$ containing $x$ is the same as $\binom{n}{k}$ since one of the $k+1$ elements is already fixed (the element $x$ ).

Notice that a subset $A \subseteq[n+1]$ either contains or doesn't contain $x$, thus we have proven the identity.

### 11.2 General Definition Of a Graph

Definition 11.2.1 (General Graph). We define a general graph $G=(V, E, d)$ as a collection of disjoint sets $V$ and $E$-called, respectively, the vertices and edges of $G$-together with a map $d$ that defines the relations of incidence (see Definition 11.2.2) between vertices and edges-that is, it defines the formation of edges between vertices.

Definition 11.2.2 (Incidence). We say that a vertex $x$ is incident with an edge $e$ if $x \in e$. Moreover if $x \in e$, we say that $e$ is an edge at $x$.

In order to ease the the way on which we talk about graphs, I'll introduce some notation that I judge will be quite appropriate to avoid confusion.

Notation 11.2.3 (Vertices and edges of a graph). Given a graph $G$, its collection of vertices and edges are denoted by, respectively, Vert( $G$ ) and Edge( $G$ ).

Notation 11.2.4 (Existence of an edge). Let $G$ be any graph and consider two vertices $x, y \in \operatorname{Vert}(G)$. We denote the relation of existence of an edge joining $x$ and $y$ by $\operatorname{Edge}(x, y)$-that is, if $\operatorname{Edge}(x, y)$ is true, there exists an edge $e \in \operatorname{Edge}(G)$ such that $x$ and $y$ are its end-vertices, otherwise, if $\operatorname{Edge}(x, y)$ is false, then there is no edge joining the two vertices.

Definition 11.2.5 (Adjacency). Given vertices $x, y \in V$, we say that $x$ and $y$ are neighbours if Edge $(x, y)$-in such case, $x$ and $y$ are said to be end-vertices of the existent edge joining them.

We say that two edges $e$ and $g$ are adjacent or neighbours if there exists a unique vertex $x$ common to both of them- $x \in e$ and $x \in g$.

A set of vertices, or edges, is said to be independent if no pair of elements is adjacent. An independent set of vertices $V$ is called stable.

Definition 11.2.6 (Diagonal set of vertices). Let $G=(V, E)$ be a graph. We define the diagonal subset of $V$ to be the collection $\Delta_{V}=\{(x, x): x \in V\}$. So that we have $V^{2} \backslash \Delta_{V}=\{(x, y): x \neq y\}$-the collection of ordered pairs of distinct vertices.

The following definition will be useful when we are talking about graphs in which there is no concept of direction, that is, an edge joining $x$ and $y$ is exactly the same as the edge joining $y$ and $x$-that is, edges have no intrinsic orientation. Graph theorists like to think about graphs as being entities that satisfy such invariance on the order of the vertices-but, when it's possible, we shall take the most general approach.

Definition 11.2.7 (Edge invariance). We define the set $\langle V\rangle^{2}$ to be the quotient

$$
\langle V\rangle^{2}=\left(V^{2} \backslash \Delta_{V}\right) /_{(x, y) \sim(y, x)} .
$$

Notation 11.2.8 (Collection of edges). Given sets $X$ and $Y$, we denote the collection of all edges of the form $x y$, where $x \in X$ and $y \in Y$, by $\operatorname{Edge}(X, Y)$. The collection of all edges that a given vertex $x$ is incident with is denoted $\operatorname{Edge}(x)$.

Definition 11.2.9 (Order and size). Let $G$ be a graph. We define the order of $G$ as $|G|=|\operatorname{Vert}(G)|$, moreover, the size of $G$ is defined as $\|G\|=|\operatorname{Edge}(G)|$.

Definition 11.2.10 (Trivial graph). A graph $G$ is said to be trivial if $|G| \leqslant 1$. In the case of the 0 order, we denote the graph as $\varnothing$.

### 11.3 The Simple Graph - Looping All Over

Definition 11.3.1 (Simple Graph). A simple graph is defined to be a graph $G=(V, E, d)$ (in the sense of Definition 11.2.1) where

$$
d: E \rightarrow V^{2} /(x, y) \sim(y, x), \text { mapping } e \stackrel{d}{\longmapsto}(x, y)=(y, x)
$$

That is, $d$ defines a reflexive and symmetric relation - that is, $\operatorname{Edge}(x, x)$ is true for all $x \in V$, and $\operatorname{Edge}(x, y)=\operatorname{Edge}(y, x)$ for all pairs $x, y \in V$.

We'll now construct a category for simple graphs using presheaves on the the following category.

Definition 11.3.2 (Walking simple graph). A walking simple graph is defined as a category S consisting of

- An object $V$, called the set of vertices - containing elements $x$.
- An object $E$, called the set of edges - containing ordered pairs $e=(x, y)$, where $x, y \in V$.
- Morphisms s, $t: E \rightarrow V$ called source and target, respectively - given an edge $e=(x, y)$ we have $e \stackrel{s}{\mapsto} x$ and $e \stackrel{t}{\mapsto} y$. We also impose that the map defined by $e \mapsto(s(e), t(e))$ in $E n d_{\text {Set }}(E)$ is injective.
- Identity morphisms $\mathrm{id}_{V}: V \rightarrow V$ and $\mathrm{id}_{E}: E \rightarrow E$ (reflexivity of vertices and edges).
- An automorphism $\sigma: E \xrightarrow{\simeq} E$ called symmetry — mapping $e=(x, y) \stackrel{\sigma}{\mapsto}(y, x)$.

Proposition 11.3.3 (Simple graph is a presheaf). Let S be a walking simple graph. A simple graph is a presheaf

$$
G: S^{\mathrm{op}} \rightarrow \text { Set. }
$$

Proof. The source and target morphisms correspond to $G(s), G(t): G(E) \rightarrow G(V)$ in such a way that, given any $e=(x, y) \in G(E), e \stackrel{G(s)}{\longmapsto} x$ and $e \stackrel{G(t)}{\longmapsto} y$ (edges have sources and targets). Moreover, since the map $e \mapsto(s(e), t(e))$ is injective, if $e, e^{\prime} \in G(E)$ are edges such that $(s(e), t(e))=\left(s\left(e^{\prime}\right), t\left(e^{\prime}\right)\right)$, then $e=e^{\prime}$ (edges are well defined). The automorphism $\sigma$ corresponds to $G(\sigma): G(E) \rightarrow G(E)$ which is again an automorphism and hence defines an equivalence between edges $(x, y)$ and $(y, x)$ - which implies that the source and target of an edge are indiscernible from each other.

Definition 11.3.4 (Morphism of simple graphs). We define morphisms of simple graphs in two equivalent ways - respectively following Definition 11.3.1 and Proposition 11.3.3:

- A morphism $G \rightarrow H$ of simple graphs $G=(V, E)$ and $H=\left(V^{\prime}, E^{\prime}\right)$ is a map $f: V \rightarrow V^{\prime}$ such that Edge $(x, y)$ implies Edge $(f(x), f(y))$.
- Let $G=(V, E, s, t)$ and $H=\left(V^{\prime}, E^{\prime}, s^{\prime}, t^{\prime}\right)$ be simple graphs. A morphism $G \rightarrow H$ is a natural transformation $\alpha: G \Rightarrow H$


Where $\alpha$ is defined by a pair of morphisms $\alpha_{V}: V \rightarrow V^{\prime}$ and $\alpha_{E}: E \rightarrow E^{\prime}$ such that the following diagrams commute

from the symmetric and reflective relations in $\operatorname{Mor}(\mathrm{S})$ we see that this is equivalent to the commutativity of the diagram ${ }^{1}$


Definition 11.3.5 (Simple graphs category). The category of simple graphs, denoted by sGraph, consists of simple graphs and morphisms between them. That is, sGraph $=$ Set ${ }^{\mathrm{S}^{\mathrm{op}}}$ - the category of presheaves on the walking simple graphs category S .

### 11.4 The Simple Loopless Graph

Sometimes (and by that I mean almost always) graph theorists like the incidence relation to be irreflexive-loops are therefore forbidden, that is, $\operatorname{Edge}(x, x)$ is always false for any $x \in V$-but this construction does not get us a good category. We are

[^17]going to reserve the word "simple graph" to mean a simple graph with loops and the word "graph" for simple graphs with no loops. The following is a formal definition of a loopless simple graph-a "graph".

Definition 11.4.1 (Graph). $G=(V, E, d)$ is said to be a graph if

$$
d: E \rightarrow\langle V\rangle^{2}, \text { mapping } e \stackrel{d}{\mapsto}(x, y)=(y, x) \text { and } x \neq y .
$$

That is, $\operatorname{Edge}(x, y)$ if and only if $\operatorname{Edge}(y, x)$, and $\operatorname{Edge}(x, x)$ is always false.
From now on we are mostly going to assume the existence of the incidence relation map $d$ (symmetric and irreflexive) and simply denote a graph $G$ by $(V, E)$-where we assume that $G$ is a simple loopless graph.

Definition 11.4.2 (Morphism of graphs). Let $G=(V, E)$ and $H=\left(V^{\prime}, E^{\prime}\right)$ be graphs. A morphism $G \rightarrow H$ is a map $\varphi: V \rightarrow V^{\prime}$ such that Edge $(x, y)$ implies Edge $(\varphi(x), \varphi(y))$ that is, adjacency of vertices need to be preserved.

Definition 11.4.3 (Graph category). We define the category of simple loopless graphs, denoted by Graph, as the category with graph objects and morphisms of graphs.

Lemma 11.4.4 (Stable preimage). Let $\varphi: G \rightarrow H$ be a morphism of graphs, then, for any $x^{\prime} \in \operatorname{Vert}(H)$, the collection $\varphi^{-1}\left(x^{\prime}\right) \subseteq \operatorname{Vert}(G)$ is stable in $G$.

Proof. Suppose, for the sake of contradiction, that there exists vertices $x, y \in \varphi^{-1}(x)$ such that $\operatorname{Edge}(x, y)$ is true in $G$. In particular, this implies that $\operatorname{Edge}(\varphi(x), \varphi(y))=$ Edge $\left(x^{\prime}, x^{\prime}\right)$ is true in $H$, which is forbidden-hence there can be no such pair of vertices in $\varphi^{-1}\left(x^{\prime}\right)$.

Definition 11.4.5 (Graph property). A class of graphs is said to be a property of graphs if it is closed up to isomorphism.

Definition 11.4.6 (Graph invariant). A map $\phi$ : Graph $\rightarrow S$ —where $S \in \operatorname{Set}$, possibly $\mathbf{R}$ for instance-is said to be a graph invariant if for all $G, G^{\prime} \in$ Graph such that $G \simeq G^{\prime}$ then

$$
\phi(G)=\phi\left(G^{\prime}\right) .
$$

Definition 11.4.7 (Union and intersections). Let $G=(V, E), G^{\prime}=\left(V^{\prime}, E^{\prime}\right) \in$ Graph. We define their union as the graph

$$
G \cup G^{\prime}=\left(V \cup V^{\prime}, E \cup E^{\prime}\right)
$$

Analogously, their intersection is defined as

$$
G \cap G^{\prime}=\left(V \cap V^{\prime}, E \cap E^{\prime}\right)
$$

Definition 11.4.8 (Complete Graph). A graph $G$ is said to be complete if for all pairs of vertices $x, y \in G$, $\operatorname{Edge}(x, y)$ is true.

Notation 11.4.9. We denote by $K^{n}$ the complete graph on $n$ vertices.
Definition 11.4.10 (Subgraph). Let $G=(V, E)$ and $G^{\prime}=\left(V^{\prime}, E^{\prime}\right)$ be graphs. $G^{\prime}$ is said to be a subgraph of $G$-denoted $G^{\prime} \subseteq G$-if $V^{\prime} \subseteq V$ and $E^{\prime} \subseteq E$. On the other hand, $G$ is said to be a supergraph of $G^{\prime}$.

Definition 11.4.11 (Induced subgraph). Let $G^{\prime}=\left(V^{\prime}, E^{\prime}\right)$ be a subgraph of $G$. $G^{\prime}$ is said to be an induced subgraph of $G$ if for all $x, y \in V$, then $\operatorname{Edge}_{G^{\prime}}(x, y)$ if $\operatorname{Edge}_{G}(x, y)$. We denote $G^{\prime}=G\left[V^{\prime}\right]$. If $G\left[V^{\prime}\right]=G$, then we say that $G^{\prime}$ is a spanning subgraph of $G$.

Definition 11.4.12 (Embedding of graphs). Let $G$ and $H$ be graphs. A morphism of graphs $t: H \hookrightarrow G$ is said to be an embedding of $H$ into $G$ if its underlying map $\operatorname{Vert}(H) \mapsto \operatorname{Vert}(G)$ is injective.

Definition 11.4.13 (Deletion vertices or edges). Let $G=(V, E)$ be a graph and $V^{\prime} \subseteq V$ a subset of the vertices of $G$. The deletion of the parts of $G$ associated with the vertices $V^{\prime}$ is defined as the graph $G\left[V \backslash V^{\prime}\right]$ —which we'll shortly identify as $G \backslash V^{\prime}$. Equivalently we define the deletion of edges $E^{\prime} \subseteq E$ as the graph $G-E^{\prime}:=\left(V, E \backslash E^{\prime}\right)$.

Notice that the non-connected vertices are not deleted from the resulting graph $G \backslash E^{\prime}$, contrary to the case of the deletion of vertices $G \backslash V^{\prime}$, where edges with endpoints out of $V^{\prime}$ where removed.

Definition 11.4.14 (Edge-maximal). Given a graph property $P$, a graph $G=(V, E)$ is said to be edge-maximal with respect to $P$ if $G \in P$ and for all $E^{\prime} \supsetneq E$ the graph $\left(V, E^{\prime}\right)$ does not have the property $P$.

Definition 11.4.15 (Clique). Let $G$ be a graph and $S \subseteq \operatorname{Vert}(G)$ be a subset with $k$ vertices. If the graph $G[S]$ is complete, we call it a $k$-clique. Trivially, $G[S] \simeq K^{k}$.

Definition 11.4.16 (Complement graph). Given a graph $G=(V, E)$, the complement of $G$ is defined as the graph $\bar{G}=\left(V,\langle V\rangle^{2} \backslash E\right)$.

Definition 11.4.17 (Line graph). Given a graph $G$, the line graph of $G$, denoted by $L(G)$, is the graph such that, if $e=(x, y), e^{\prime}=(y, z) \in \operatorname{Edge}(G)$ are adjacent edges then $\operatorname{Edge}_{L(G)}(x, z)$-that is $L(G)=(\operatorname{Vert}(G), E)$ where

$$
E=\left\{(x, z) \in\langle V\rangle^{2}: \operatorname{Edge}_{G}(x, y) \text { and } \operatorname{Edge}_{G}(y, z) \text { for some } y \in \operatorname{Vert}(G)\right\}
$$

Definition 11.4.18 (Join). Let $G=(V, E), G^{\prime}=\left(V^{\prime}, E^{\prime}\right) \in$ Graph be disjoint- $G \cap G^{\prime}=\varnothing$. We define their join as the graph $G * G^{\prime}$ with vertices $V \cup V^{\prime}$ and edges $E \cup E^{\prime} \cup E^{*}$, where $E^{*}$ is defined as $E^{*}=\left\{(x, y): x \in G, y \in G^{\prime}\right\}$-that is, the collection of all edges connecting the vertices of $G$ and $G^{\prime}$.

The join definition fails to be a coproduct in the category of graphs exactly because of the addition of the $E^{*}$ edges-the edges between elements of $G$ with $G^{\prime}$ in $G * G^{\prime}$ cannot be ensured to be preserved by a morphism $G * G^{\prime} \rightarrow H$ as described in Proposition 11.6.4.

## Vertex Degree

Notation 11.4.19 (Neighbourhood). Given a graph $G$, we denote by $N(v)$ the collection of all neighbours of $v \in G$.

Definition 11.4.20 (Degree). Let $G$ be a graph, we define the degree (or valency) of a vertex $v \in G$ as $\operatorname{deg} v=|N(v)|$. If $G$ is a finite graph,

- The minimum degree of $G$ is defined as $\delta(G)=\min _{v \in \operatorname{Vert}(G)} \operatorname{deg} v$.
- The maximum degree of $G$ is defined as $\Delta(G)=\max _{v \in \operatorname{Vert}(G)} \operatorname{deg} v$.
- The average degree of $G$ is defined as

$$
\operatorname{deg} G=\frac{1}{|G|} \sum_{v \in \operatorname{Vert}(G)} \operatorname{deg} v
$$

- The edge-vertex ratio of $G$ is $\epsilon(G)=\frac{\|G\|}{|G|}$. Since an edge is composed of two vertices, the sum $\sum_{v \in \operatorname{Vert}(G)} \operatorname{deg} v$ counts every edge exactly twice, so that $\epsilon(G)=$ $\frac{1}{2} \sum_{v \in \operatorname{Vert}(G)} \operatorname{deg} v$. Thus we arrive in the relation $\epsilon(G)=\frac{1}{2} \operatorname{deg} G$.

Lemma 11.4.21. The number of vertices of odd degree in a graph is even.
Proof. Let $G$ be a graph. Since $\|G\| \in \mathbf{N}$ and $\|G\|=\frac{1}{2} \sum_{v \in \operatorname{Vert}(G)} \operatorname{deg} v$, then

$$
\sum_{v \in \operatorname{Vert}(G)} \operatorname{deg} v
$$

is even. If the number of odd degree vertices where odd, then the sum of their degrees would also be odd. Since the sum of even degree vertices is always even, it is necessary for the number of odd degree vertices to be even.

Proposition 11.4.22 (Edge-dense subgraph). Let $G$ be a finite graph with size $\|G\| \geqslant 1$. Then there exists a subgraph $H \subseteq G$ such that

$$
\delta(H)>\epsilon(H) \geqslant \epsilon(G)
$$

Proof. We'll construct the subgraph $H$ by means of a chain of consequent single vertex deletion-as in Definition 11.4.13. Let $H_{0}=G$ Notice that, if we want to have the edge-vertex ratio unaltered (or increased) by the deletion, we got to choose a vertex $v \in H_{0}$-if existent-such that the graph $H_{1}=H_{0} \backslash v$ has $\epsilon\left(H_{1}\right) \geqslant \epsilon\left(H_{0}\right)$, that is, $\operatorname{deg} v$ has to satisfy

$$
\epsilon\left(H_{1}\right)=\frac{\left\|H_{0}\right\|-\operatorname{deg} v}{\left|H_{0}\right|-1} \geqslant \frac{\left\|H_{0}\right\|}{\left|H_{0}\right|}=\epsilon\left(H_{0}\right) .
$$

Solving the above equation for the degree of $v$, we find that $\operatorname{deg} v \leqslant \epsilon\left(H_{0}\right)$. Since $\delta\left(H_{0}\right) \leqslant \operatorname{deg} H_{0} \leqslant \Delta\left(H_{0}\right)$ and $\epsilon\left(H_{0}\right) \leqslant \operatorname{deg} H_{0}$, then $\operatorname{deg} v \leqslant \Delta\left(H_{0}\right)$-that is, $\delta\left(H_{1}\right) \leqslant$ $\delta\left(H_{0}\right)$. If such vertex doesn't exists in $\operatorname{Vert}\left(H_{0}\right)$, we terminate our algorithm and
find $H=H_{0}=G$. Otherwise, we continue recursively, generating a finite chain of subgraphs-the finiteness of the chain comes from the fact that $G$ itself is finite

$$
G=H_{0} \supset H_{1} \supset H_{2} \supset \cdots \supset H_{n}=H
$$

for some $n \in \mathbf{N}$. Where $\epsilon\left(H_{j+1}\right) \geqslant \epsilon\left(H_{j}\right)$ and $\delta\left(H_{j+1}\right) \geqslant \delta\left(H_{j}\right)$ for all $0 \leqslant j<n$. At some point the chain will terminate into the subgraph $H$. Moreover, since the recursion terminates in $H \neq \varnothing$, then necessarily $\operatorname{deg} v>\epsilon(H)$ for all $v \in H$-in particular, this implies in $\delta(H)>\epsilon(H)$.

Definition 11.4.23 (Regular graph). A graph $G$ is said to be $k$-regular-for some $k \in$ $\mathbf{N}$-if $\operatorname{deg} v=k$ for all $v \in G$.

### 11.5 Path to Glory

Definition 11.5.1 (Walk). A walk on a graph $G$ is defined to be a subgraph $W \subseteq G-$ assume $|W|=n+1$ for some $n \in \mathbf{N}$-such that there exists a surjective map

$$
\ell:[n] \rightarrow \operatorname{Vert}(W)
$$

satisfying the condition of $\operatorname{Edge}_{G}(\ell(k-1), \ell(k))$ for all $1<k<n$. We'll generally assume this underlying surjective vertex-labelling map and denote $W=(\ell(1), \ldots, \ell(n))=$ $\left(v_{1}, \ldots, v_{n}\right)$. The walk $W$ is said to be closed if $v_{1}=v_{n}$.

Definition 11.5.2 (Length). We define the length of a walk $W$ to be its size $\|W\|$.
Definition 11.5.3 (Walk operations). Given a path $W=\left(v_{0}, \ldots, v_{n}\right)$ we can define the following:

- (Subwalk) Let $v_{j} \in \operatorname{Vert}(W)$ be any vertex. We define a subwalk of $W$ restricted to $v_{j}$ to be one of the following subgraphs of $W$ : $v_{j} W=\left(v_{j}, v_{j+1}, \ldots, v_{n}\right)$ and $W v_{j}=\left(v_{0}, \ldots, v_{j-1}, v_{j}\right)$. We can also define excluding subwalk induced by $v_{j}$ to be $\breve{v}_{j} W=\left(v_{j+1}, \ldots, v_{n}\right)$ and $W \breve{v}_{j}=\left(v_{0}, \ldots, v_{j-1}\right)$.
- (Inner walk) The inner walk of $W$ is given by $\breve{W}=\left(v_{1}, \ldots, v_{n-1}\right)$.
- (Walk concatenation) Let $Q=\left(w_{0}, \ldots, w_{m}\right)$ be a walk. If $P$ and $Q$ have coinciding endings, say $v_{n}=w_{0}=x$, we can define the concatenation of $P$ with $Q$ to be the walk $P Q=\left(v_{0}, \ldots, v_{n-1}, x, w_{1}, \ldots, w_{m}\right)$.

Definition 11.5.4 (Path). We define a path to be walk with no repeating vertices. A path graph $P_{n}$ is a graph of order $n$ and size $n-1$ such that there exists a bijective map

$$
\ell_{n}:[n] \xrightarrow{\simeq} \operatorname{Vert}\left(P_{n}\right)
$$

for which $\operatorname{Edge}_{P_{n}}\left(\ell_{n}(k-1), \ell_{n}(k)\right)$ for all $1<k<n$. We'll usually assume the existence of the underlying vertex-labelling bijection $\ell_{n}$. This way we can naturally write the path $P_{n}$ as a unique sequence of vertices $\left(v_{1}, \ldots, v_{n}\right)$ up to isomorphism of graphs.

Given a graph $G$, we define a path of size $n$ on $G$ to be the induced subgraph of $G$ given by $G\left[\iota\left(P_{n}\right)\right]$, where $\iota: P_{n} \hookrightarrow G$ is an embedding of graphs—see Definition 11.4.12. The underlying labelling is now given by the composition $\iota \ell_{n}:[n] \rightarrow \operatorname{Vert}(G)$.

Since a path is just a special type of walk, every operation described on Definition 11.5.3 is carried over to the path graphs.

Proposition 11.5.5 (Transforming paths into walks). A map $\phi: P_{n} \rightarrow G$ is a morphism of graphs if and only if the sequence $\left(\phi\left(v_{0}\right), \ldots, \phi\left(v_{n-1}\right)\right)$ is a walk on $G$.

Proof. Throughout, assume $1<j<n$. If $\phi$ is a morphism of graphs then $\operatorname{Edge}_{P_{n}}\left(v_{j-1}, v_{j}\right)$ implies $\operatorname{Edge}_{G}\left(\phi\left(v_{j-1}\right), \phi\left(v_{j}\right)\right)$ thus inducing a walk on $G$. On the other hand, if

$$
\left(\phi\left(v_{1}\right), \ldots, \phi\left(v_{n}\right)\right)
$$

is a walk on $G$, then since the only edges on $P_{n}$ occur between consequent vertices $v_{j-1}$ and $v_{j}$, we have that $\phi$ indeed preserves the adjacency structure of the path $P_{n}$, that is, $\operatorname{Edge}_{P_{n}}\left(v_{j-1}, v_{j}\right)$ implies $\operatorname{Edge}_{G}\left(\phi\left(v_{j-1}\right), \phi\left(v_{j}\right)\right)$-hence $\phi$ is a morphism.

Lemma 11.5.6 (Paths on walks). Let $W$ be a walk. If $x, y \in \operatorname{Vert}(W)$, then there exists a path $P \subseteq W$ such that $x, y \in \operatorname{Vert}(P)$.

Proof. Assume $W$ is a $k$-walk and contains $x$ and $y$ at its end-vertices-if not, we would be analogously analysing the subwalk $x W y$. Let $\ell:[k] \rightarrow \operatorname{Vert}(W)$ be a surjective labelling on $W$-since $x$ and $y$ are end-vertices, we have

$$
\begin{equation*}
\ell(1)=x \text { and } \ell(k)=y . \tag{11.1}
\end{equation*}
$$

Let $C$ be the collection of all cycles on $W$-refer to Definition 11.5.10. Define the equivalence relation $\sim$ on the vertices $\operatorname{Vert}(W)$ to be so that $v \sim u$ if and only if $v, u \in \operatorname{Vert}(C)$ for some $C \in C$.

The quotient graph $W / \sim$ (see Definition 11.6.1) is a path joining the vertex classes $\bar{x}$ and $\bar{y}$. To see that, we can first construct an equivalence relation $\sim_{\ell}$ on the indexing set [k], defined as follows: $i \sim_{\ell} j$ if and only if $\ell(i) \sim \ell(j)$ —clearly $[k] / \sim_{\ell}$ is an indexing set for the vertices of the quotient graph. Lets assume that $[k] / \sim_{\ell}=\{\overline{1}, \ldots, \bar{m}\}$. Let $\ell^{\prime}:[k] / \sim_{\ell} \rightarrow \operatorname{Vert}(W / \sim)$ be a labelling on $W / \sim \operatorname{defined}$ as $\ell^{\prime}(\bar{j})=\overline{\ell(j)}$. Using Eq. (11.1), the sequence of vertices $\left(\ell^{\prime}(\overline{1}), \ldots, \ell^{\prime}(\bar{m})\right)$ is such that $\ell^{\prime}(\overline{1})=\bar{x}$ and $\ell^{\prime}(\bar{m})=\bar{y}$. Also, $\operatorname{Edge}_{W / \sim}\left(\ell^{\prime}(\bar{j}), \ell^{\prime}(\overline{j+1})\right)=\operatorname{Edge}_{W}(\ell(j), \ell(j+1))$ is always true (because $W$ is a walk). So far, we've shown that $W / \sim$ is a walk between $\bar{x}$ and $\bar{y}$-we now need to show that no vertex is repeated in the given labelled sequence, which is actually trivial. Suppose that $\ell^{\prime}(\bar{i})=\ell^{\prime}(\bar{j})$, then from definition of $\ell^{\prime}$ we have $\overline{\ell(i)}=\overline{\ell(j)}$, which implies that $\ell(i) \sim \ell(j)$ and hence $i \sim_{\ell} j$-thus $\bar{i}=\bar{j}$. This shows that

$$
W / \sim=\left(\bar{x}, \ell^{\prime}(\overline{1}), \ldots, \ell^{\prime}(\overline{m-1}), \bar{y}\right)
$$

is indeed a path joining $\bar{x}$ and $\bar{y}$.
For the last part we just need to consider the embedding of graphs $t: W / \sim \hookrightarrow W$ such that $\operatorname{Edge}_{W / \sim}(\bar{v}, \bar{u})$ implies $\operatorname{Edge}_{W}(\iota(\bar{v}), \iota(\bar{u}))$. The sequence of vertices

$$
\iota(W / \sim):=\left(\iota(\bar{x}), \iota \ell^{\prime}(\overline{1}), \ldots, \iota \ell^{\prime}(\overline{m-1}), \iota(\bar{y})\right) \subseteq W
$$

is clearly a path on $W$ joining the vertices $\iota(\bar{x})=x$ and $\iota(\bar{y})=y$. Hence, $P=W[\operatorname{im} \iota]$ gives us the wanted path-finally the proposition is proved.

The last proof may have got really clumsy at some points, so here goes an example of the basic operations we've developed and used throughout the proof. Let $W=$ $\left(x, v_{1}, \ldots, v_{12}, y\right)$ be a walk, visually given by-where $v_{5}=v_{2}$ and $v_{11}=v_{6}$


Now, the process of taking the quotient of the graph amounts to the identification of the cycles

$$
C=\left\{W\left[\left\{v_{2}, v_{3}, v_{4}, v_{5}\right\}\right], W\left[\left\{v_{6}, v_{7}, v_{8}, v_{9}, v_{10}, v_{11}\right\}\right]\right\}
$$

The cycles are respectively reduced to classes $\bar{v}_{2}$ and $\bar{v}_{3}$, both elements of $\operatorname{Vert}(W) / \sim$. The quotient graph $W / \sim$ can be visually depicted as follows


Now for the embedding of graphs $t: W / \sim \hookrightarrow W$, we can view $P=W[\operatorname{im} t]$ as the subgraph consisting of the red edges and their end-vertices-that is, the collection of vertices $\left\{x, v_{1}, v_{2}, v_{6}, v_{12}, y\right\}$-which can be visualised as follows


Corollary 11.5.7 (Distances \& morphisms of graphs). Consider graphs $G$ and H. If $\phi: G \rightarrow H$ is a morphism of graphs then, for all $x, y \in G$, we have

$$
d_{H}(\phi(x), \phi(y)) \leqslant d_{G}(x, y)
$$

Proof. Assume $\phi$ is a morphism. Suppose $x, y \in G$ are separated by a finite distanceif not, the proposition follows trivially. Let $P_{k}$ be a minimal path on $G$ joining $x$ and $y$. Since $\phi$ is a morphism, the image $\phi\left(P_{k}\right) \subseteq H$ is a walk where $\phi(x), \phi(y) \in P_{k}$-this follows directly from Proposition 11.5.5. Now, using Lemma 11.5.6 on $\phi\left(P_{k}\right)$ and the vertices $\phi(x)$ and $\phi(y)$, it follows that there exists a path $P \subseteq \phi\left(P_{k}\right)$ joining $\phi(x)$ and $\phi(y)$. Moreover, clearly $\|P\| \leqslant\left\|\phi\left(P_{k}\right)\right\| \leqslant k$, thus $d_{H}(\phi(x), \phi(y)) \leqslant d_{G}(x, y)$.

Definition 11.5.8 (Path between sets). Let $A$ and $B$ be disjoint sets of vertices. $P=$ $\left(v_{0}, \ldots, v_{n}\right)$ is said to be an $A-B$ path if $v_{0} \in A$ and $v_{n} \in B$.

Definition 11.5.9 (Independent paths). Two paths are said to be independent if their inner path share no vertex.

## Cycles

Definition 11.5.10. A cycle is a closed walk with size at least 3. A $k$-cycle is a cycle $C^{k}$ whose length is $k$.

A $k$-cycle on a graph $G$ is the induced subgraph of $G$ given by $G\left[\iota\left(C^{k}\right)\right]$ such that $\iota: C^{k} \rightarrow G$ is an embedding of graphs.

Definition 11.5.11 (Chords). Let $G$ be a graph and $C$ be a cycle on $G$. Given $x, y \in C$, a chord is an edge $(x, y) \in \operatorname{Edge}(G)$ such that $\operatorname{Edge}_{C}(x, y)$ is false.

Definition 11.5.12 (Induced cycle). An induced cycle on a graph is a cycle with no chords.

Definition 11.5.13 (Distance). Let $G$ be a graph and $x, y \in \operatorname{Vert}(G)$ be any vertices. Let $\mathcal{P}$ be the collection of all paths in $G$ containing the vertices $x$ and $y$. If $\mathcal{P}$ is non-empty, we define the distance of $x$ and $y$ on $G$ to be the minimum length of the subpaths linking $x$ to $y$, in other words

$$
d_{G}(x, y)=\min _{P \in \mathcal{P}}\|x P y\| .
$$

On the other hand, if $\mathcal{P}$ is empty, then $d_{G}(x, y)=\infty$.
Definition 11.5.14 (Miscellaneous definitions). We define the following:
(a). (Girth) The girth of a graph $G$ is the minimum length of a cycle on $G$. In other words, let $C$ be the collection of all cycles on $G$, the girth of $G$ is

$$
g(G)=\min _{C \in C}\|C\| .
$$

(b). (Circumference) The circumference of a graph $G$ is the maximal length of a cycle on $G$. In other words, let $C$ be the collection of all cycles on $G$. If $C$ is non-empty, we set

$$
\operatorname{circ}(G)=\max _{C \in C}\|C\|
$$

(c). (Diameter) The diameter of a graph $G$ is the maximal distance between two vertices of $G$, that is

$$
\operatorname{diam}(G)=\max _{x, y \in \operatorname{Vert}(G)} d_{G}(x, y)
$$

Proposition 11.5.15. Let $G$ be a graph such that $\delta(G) \geqslant 2$. Then there exists a path on $G$ with length $\delta(G)$ and a cycle on $G$ with length at least $\delta(G)+1$.

### 11.6 Universal Properties of Graph

## Quotients

Definition 11.6.1 (Quotient). Let $G$ be a graph and $\sim$ be an equivalence relation on the collection of vertices $\operatorname{Vert}(G)$. We define the quotient graph $G / \sim$ to be the graph whose vertices are the vertex classes $\operatorname{Vert}(G) / \sim$ and whose edges are such that Edge $_{G / \sim}([x],[y])$ if and only if $E^{-1 g e}{ }_{G}(x, y)$.

Proposition 11.6.2 (Universal property of quotients). The quotient graph is a quotient in the category Graph . In other words, let $G$ be a graph and $\sim$ be an equivalence relation on Vert(G). Consider any graph $H$ together with a morphism of graphs $\psi: G \rightarrow H$. Then there exists a unique morphism of graphs $\phi: G / \sim \rightarrow H$ such that the following diagram commutes

where $\pi$ is the naturally defined projection.

Proof. First we show that $\phi$ is indeed unique. Suppose $\phi_{1}$ and $\phi_{2}$ are both morphisms that satisfy the commutativity of the diagram. Given any $x \in \operatorname{Vert}(G)$ we have $\psi(x)=$ $\phi_{1}([x])=\phi_{2}([x])$ and since $\pi(\operatorname{Vert}(G))=\operatorname{Vert}(G / \sim)$ (surjective property) then $\phi_{1}$ and $\phi_{2}$ have equal images throughout their whole domain - implying $\phi_{1}=\phi_{2}$. To show that $\phi$ is a morphism of graphs, it is sufficient to consider any $x, y \in G$ such that $\operatorname{Edge}_{G}(x, y): \psi$ being a morphism implies $\operatorname{Edge}_{G}(\psi(x), \psi(y))$, then - since $\phi([x])=\psi(x)$ and $\phi([y])=\psi(y)$ - we get that $\operatorname{Edge}_{G}(\phi([x]), \phi([y]))$ is true.

## Coproducts

Definition 11.6.3 (Coproduct). Let $G=(V, E)$ and $H=\left(V^{\prime}, E^{\prime}\right)$ be graphs, we define their coproduct $G \oplus H$ to be the disjoint union of vertices and edges of the original graphs - that is, $G \oplus H=\left(V \amalg V^{\prime}, E \amalg E^{\prime}\right)^{2}$.

Proposition 11.6.4 (Universal property of coproducts). Given graphs $G, H \in$ Graph, the graph $G \oplus H$ is a coproduct of $G$ and $H$ in the category of simple loopless graphs. That is, given any graph $W \in$ Graph and graph morphisms $f: G \rightarrow W$ and $g: H \rightarrow W$, there exists a unique graph morphism $\phi: G \oplus H \rightarrow W$ such that the following diagram

[^18]commutes


Where the inclusion morphisms $\iota_{G}$ and $\iota_{H}$ are naturally defined.
Proof. Notice that $\phi$ is defined by a map of vertices $\operatorname{Vert}(G) \cup \operatorname{Vert}(H) \rightarrow \operatorname{Vert}(W)$ with the restriction that $\operatorname{Edge}_{G \oplus H}(x, y)$ implies $\operatorname{Edge}_{W}(\phi(x), \phi(y))$, we'll first prove that $f$ together with $g$ completely identify $\phi$. Let $v \in G \oplus H$ be any vertex. If $v \in \operatorname{Vert}(G)$, suppose $f(v)=h$ then necessarily $\phi \iota_{G}(v)=h$. On the other hand, if $v \in \operatorname{Vert}(H)$ and if $g(v)=h^{\prime}$ then $\phi \iota_{H}(v)=h^{\prime}$. Since $\operatorname{cod} \phi=\operatorname{Vert}(G) \amalg \operatorname{Vert}(H)$ and we have $\iota_{G}(G)=\operatorname{Vert}(G)$ and $\iota_{H}(H)=\operatorname{Vert}(H)$, this shows that $f$ and $g$ completely determine the image of $\phi$ - hence $\phi$ is uniquely defined.

Finally we show that $\phi$ is indeed a morphism of graphs. Given $x, y \in \operatorname{Vert}(G)$, suppose $\operatorname{Edge}_{G}(x, y)$, then $\operatorname{Edge}_{W}(f(x), f(y))=\operatorname{Edge}_{W}(\phi(x), \phi(y))-$ since $\phi(x)=$ $f(x)$ and $\phi(y)=f(y)$. The case $x, y \in \operatorname{Vert}(H)$ is completely analogous and we'll therefore omit for the sake of brevity. We don't need to inspect the case where $x \in$ $\operatorname{Vert}(G)$ and $y \in \operatorname{Vert}(H)$ since $\operatorname{Edge}_{G \oplus H}(x, y)$ is always false in such instance. Thus $\phi$ is a morphism of graphs.

## Products

Definition 11.6.5 (Kronecker product). Let $G, H \in$ Graph. We define the Kronecker product of $G$ and $H$ to be the graph $G \otimes H=(V, E)$ whose vertices are $V=\operatorname{Vert}(G) \times$ $\operatorname{Vert}(H)$ and edges defined by $\operatorname{Edge}_{G \otimes H}\left(v \otimes h, v^{\prime} \otimes h^{\prime}\right)$ if and only if $\operatorname{Edge}_{G}\left(v, v^{\prime}\right)$ and Edge $_{H}\left(h, h^{\prime}\right)$.

Proposition 11.6.6 (Products). The Kronecker product as defined above is a product in the category of simple loopless graphs Graph. That is, given graphs $G, H, W \in \operatorname{Graph}$ and graph morphisms $f: W \rightarrow G$ and $g: W \rightarrow H$, there exists a unique morphism of graphs $\phi: W \rightarrow G \otimes H$ such that the following diagram commutes


Where $\pi_{G}$ and $\pi_{H}$ are the naturally defined projection morphisms.

Proof. Consider the morphism $\phi: W \rightarrow G \otimes H$ defined by the mapping $w \stackrel{\phi}{\mapsto} f(w) \otimes$ $g(w)$. Let $w, w^{\prime} \in \operatorname{Vert}(W)$ be any vertices such that $\operatorname{Edge}_{W}\left(w, w^{\prime}\right)$, since $f$ and $g$ are graph morphisms, then $\operatorname{Edge}_{G}\left(f(w), f\left(w^{\prime}\right)\right)$ and $\operatorname{Edge}_{H}\left(g(w), g\left(w^{\prime}\right)\right)$. Since $\phi(w)=$ $f(w) \otimes g(w)$ and $\phi\left(w^{\prime}\right)=f\left(w^{\prime}\right) \otimes g\left(w^{\prime}\right)$ it follows from the construction of the Kronecker product that $\mathrm{Edge}_{G \otimes H}\left(\phi(w), \phi\left(w^{\prime}\right)\right)$. This shows that $\phi$ is a graph morphism.

We now inspect its uniqueness. Let $\phi, \phi^{\prime} \in \operatorname{Mor}_{G r a p h}(W, G \otimes H)$ be morphisms satisfying the commutativity of the diagram. Then $f=\pi_{G} \phi=\pi_{G} \phi^{\prime}$ and $g=\pi_{H} \phi=$ $\pi_{H} \phi^{\prime}$, thus, given any $w \in W$ we have $\pi_{G} \phi(w)=\pi_{G} \phi^{\prime}(w)$ and $\pi_{H} \phi(w)=\pi_{H} \phi^{\prime}(w)$, which implies in $\phi(w)=\phi^{\prime}(w)$. Therefore $\phi$ is unique.

### 11.7 Ramsey and Schur theorem

Theorem 11.7.1 (Ramsey theorem on graphs). Given $r \in \mathbf{N} \geqslant 1$ we have that the colouring function $C:\binom{\mathbf{N}}{2} \rightarrow[r]$ is such that there exists an infinite monochromatic subset $A \subseteq \mathbf{N}$, that is, for all $a, b \in A$, we have $C(a b)=i \in[r]$.
Proof. Firstly, define the sets $\Gamma_{k}(x):=\{y \in \mathbf{N}: \mathcal{C}(x y)=k \in[r]\}$. We know that for any element $x \in \mathbf{N}$ the colouring function $C$ provides at most $r$ of such sets. Notice that for every vertex $v_{i} \in \mathbf{N}$ there are infinitely many edges connecting to other vertices of $\mathbf{N}$ this way we conclude that since $[r]$ partitions $\mathbf{N}$ only into a finite number of subsets, by the pigeonhole principle, we can conclude that there is one set, which we'll denote by $\Gamma_{k_{i}}$, that is infinite.

Let now the sequence $\left(v_{n}\right)_{n \in \mathbf{N}}$ such that for all $n \in \mathbf{N}, v_{n} \in \Gamma_{k_{n-1}}\left(v_{n-1}\right)$ for some colour $k_{n-1} \in[r]$. A trivial result of such construction is that given any $n \in N$ we have $C\left(v_{n} v_{n-1}\right)=k_{n-1}$. In fact we can extend such result by simply recalling that since we are always taking the infinite sets, surely we'll get the sequence

$$
\left(\Gamma_{k_{n}}\left(v_{n}\right)\right)_{n \in \mathbf{N}} \text { such that } \mathbf{N} \supseteq \Gamma_{k_{1}}\left(v_{1}\right) \supseteq \Gamma_{k_{2}}\left(v_{2}\right) \supseteq \ldots
$$

What this means is that given any pair $\left\{v_{i}, v_{j}\right\}$ of elements of such sequence, we have that $C\left(v_{i}, v_{j}\right)=k_{\min \{i, j\}}$. Define now the sequence $\left(k_{\min \{i, j\}}\right)_{\{i, j\} \in\binom{N}{2}}$, since this is a sequence with infinitely many elements, we can say that the finite colouring [ $r$ ] induces a finite partitioning of such sequence. By the pigeonhole principle we conclude finally that there exists an infinite monochromatic subsequence. With this we conclude the proof since this construction is obtained directly from $\left(v_{n}\right)_{n \in \mathbf{N}}$.

We actually didn't stated the general version of Ramsey theorem, which extends for the colouring of the set $\binom{X}{k}$ where $X$ is a countably infinite poset and $k \in \mathbf{N}$.
Definition 11.7.2 (Hypergraph). An ordered pair $H=(V, E)$ of sets, where $V$ are the vertices and $E$ the hyperedges, is called an hypergraph if $E \subseteq\binom{V}{k}$, thus an edge connect two or more vertices.

Theorem 11.7.3 (Ramsey theorem on Hypergraphs). Let $X$ a countably infinite poset. Then, for all $k, r \in \mathbf{N}$, the colouring $C:\binom{X}{k} \rightarrow[r]$ is such that there exists an infinite subset $A \subseteq X$ such that, for all $a, b \in A$, we have $C(a b)=i \in[r]$.

Proof. We proceed via induction on $k$. For the case $k=1$ we have that $\binom{X}{1}=X$ and thus the colouring $C$ induces a partition of the infinite set $X$ into a finite number of subsets. By the pigeonhole principle we conclude that there is one of such subsets that is infinite and also monochromatic.

For the inductive step, let $\theta \in X$ be the smallest element of $X$ (we can do such a thing since we said $X$ was a poset), let the colouring $C:\binom{X}{k+1} \rightarrow[r]$ and another colouring

$$
C_{0}:\binom{X \backslash\{\theta\}}{k} \longrightarrow[r] \text { such that } C_{0}(E):=C(E \cup\{\theta\}), \text { for all } E \in\binom{X}{k}
$$

By the inductive hypothesis we say that there exists an infinite subset $X_{0} \subseteq X$ such that, for all $E_{0} \in\binom{X_{0}}{k}$, we have $C_{0}\left(E_{0}\right)=c_{0} \in[r]$.

Now let $X_{0}^{\prime}:=\left\{x \in X_{0}: \theta<x\right\}$ and define $x_{1}:=\min \left(X_{0}\right)$. As before, set the colouring $C_{1}:\binom{X_{0}^{\prime}}{k} \rightarrow[r]$ such that $C_{1}(E):=C\left(E \cup\left\{x_{1}\right\}\right)$. Since $X_{0}^{\prime}$ is an infinite set by construction, we know from the pigeonhole principle that there exists an infinite subset $X_{1} \subseteq X_{0}^{\prime}$ such that, for all $E_{1} \in\binom{X_{1}}{k}$, we have $C_{1}\left(E_{1}\right)=c_{1} \in[r]$.

We continue constructing a sequence of elements $\left(\theta, x_{1}, x_{2}, \ldots\right)$ in such a way that we have a corresponding sequence $\left(c_{0}, c_{1}, c_{2}, \ldots\right)$ of colours. Note now that we can construct a new and final colouring function

$$
C^{\prime}:\left\{\theta, x_{1}, x_{2}, \ldots\right\} \rightarrow[r] \text { such that } C^{\prime}\left(x_{j}\right):=c_{j} \in[r] .
$$

There are countably infinite elements on the domain and only finite colours, the partitioning is such that the pigeonhole principle is applicable and thus there is a infinite subset $A \subseteq\left\{\theta, x_{1}, x_{2}, \ldots\right\} \subseteq X$ such that for all $x \in A$ we have $C^{\prime}(x)=c \in[r]$. This concludes the proof.

## Part IV

## Classical Topology

## Chapter 12

## Topological Spaces

### 12.1 Topology

Definition 12.1.1 (Topology). Let $X$ be a set and $\tau \subseteq 2^{X}$. We say that $\tau$ is a topology for $X$ if the following properties are satisfied
(T1) $X, \varnothing \in \tau$.
(T2) The arbitrary union of elements of $\tau$ is an element of $\tau$.
(T3) The finite intersection of elements of $\tau$ is an element of $\tau$.
The elements of the topology $\tau$ are called open sets of $X$.
Example 12.1.2. We proceed by listing some examples of topologies that are somewhat interesting, they are included here in order to familiarize the reader with the possible constructions for the topology on a given set $X$ :

- $\tau_{1}=\{U \subseteq X: U=\varnothing$ or $X \backslash U$ is finite $\}$ is the cofinite topology on $X$.
- $\tau_{2}=\{U \subseteq X: U=\varnothing$ or $X \backslash U$ is countable $\}$ is the cocountable topology on $X$.
- Let $p \in X$, then $\tau_{3}=\{U \subseteq X: U=\varnothing$ or $p \in U\}$ is the particular point topology on X.
- Let $p \in X$, then $\tau_{4}=\{U \subseteq X: U=X$ or $p \notin U\}$ is the excluded point topology on X.
- The collection $2^{X}$ forms what is called the discrete topology on $X$.

Definition 12.1.3 (Comparing topologies). Let $X$ be a set, and $\tau$ and $\tau^{\prime}$ be topologies on $X$. If $\tau^{\prime} \supseteq \tau$, then we say that

- $\tau^{\prime}$ is finer than $\tau$. Moreover, if $\tau$ is strictly contained in $\tau^{\prime}$, we say that $\tau^{\prime}$ is strictly finer than $\tau$
- $\tau$ is coarser than $\tau^{\prime}$. Moreover, if $\tau$ is strictly contained in $\tau^{\prime}$, we say that $\tau^{\prime}$ is strictly coarser than $\tau$

In general, we say that two topologies are comparable if either of them contains the other.

## Topological Basis

Definition 12.1.4 (Basis). Let $X$ be a topological space. A collection $\mathcal{B} \subseteq 2^{X}$ is said to be a basis for the topology of $X$ if it satisfies the following
(B1) Every element of $\mathcal{B}$ is an open set of $X$.
(B2) Every open subset $U \subseteq X$ can be written as a union of elements of $B$, that is, exists $\left\{B_{i}\right\}_{i \in I} \subseteq \mathcal{B}$ for which $U=\bigcup_{i \in I} B_{i}$.

Proposition 12.1.5 (Necessary and sufficient condition for a basis). Let $X$ be a set and $\mathcal{B} \subseteq 2^{X}$. Then $\mathcal{B}$ is a basis for some topology of $X$ if and only if it satisfies

B1 $X=\bigcup_{B \in \mathcal{B}} B$.
B2 If $x \in A \cap B$, where $A, B \in \mathcal{B}$, then there exists $C \in \mathcal{B}$ such that $x \in C \subseteq A \cap B$.
Proof. Let $\mathcal{B}$ be a basis for the space $X$. Let $p \in X$ be any point and let $U \subseteq X$ be a neighbourhood of $x$. From the definition of a basis, there exists a subcollection of sets such that their union equals $U$, which implies in the existence of $B \in \mathcal{B}$ such that $x \in B$. Moreover, since $B \subseteq X$ for all $B \in \mathcal{B}$, it follows that $X=\bigcup_{B \in \mathcal{B}} B$. Let $A, B \subseteq \mathcal{B}$ be any intersecting sets of $\mathcal{B}$ and take any point $p \in A \cap B$. Notice that $A \subseteq B \subseteq X$ is open, hence there exists a subcollection of open sets of the basis $\mathcal{B}$ whose union equals $A \cap B$. It is immediate that there exist $C \in \mathcal{B}$ such that $p \in C$ and necessarily $C \subseteq A \cap B$.

Let $\mathcal{B} \subseteq 2^{X}$ satisfying both conditions specified above. We first show that $\mathcal{B}$ is a collection of open sets. Let $\tau$ be the collection of all possible unions of sets of $\mathcal{B}$. From the first property, $X \in \tau$ and clearly $\varnothing \in \tau$ - satisfying Item 1. From construction, unions of sets in $\tau$ are unions of unions of sets of $\mathcal{B}$, which is certainly contained in $\tau$ - hence the collection satisfies Item 2. Let $T, T^{\prime} \in \tau$ be any intersecting sets and, for every $p \in T \cap T^{\prime}$, choose any $B, B^{\prime} \subseteq \mathcal{B}$ such that $p \in B \subseteq T$ and $p \in B^{\prime} \subseteq T^{\prime}-$ which are ensured to exist. From the second property of $\mathcal{B}$, there exists $C \in \mathcal{B}$ such that $p \in C \subseteq B \cap B^{\prime}$, thus $C \subseteq B \cap B^{\prime}$ and, in particular $C \subseteq T \cap T^{\prime}$. We can see that $T \cap T^{\prime}$ is again the union of a collection of elements of $\mathcal{B}$, hence $\tau$ is closed under finite intersections, satisfying Item 1 . We can now finally conclude that $\tau$ is a topology on $X$ and hence $\mathcal{B}$ is composed of open sets of $X-$ and, even better than that, $\tau$ is the unique topology generated by $\mathcal{B}$.

Definition 12.1.6 (Subbase). Let $(X, \tau)$ be a topological space. A collection $\mathcal{S} \subseteq \tau$ is called a subbase for $(X, \tau)$ if the collection of all finite intersections $U_{1} \cap \cdots \cap U_{n}$, where $U_{i} \in \mathcal{S}$, is a base for $(X, \tau)$.

Definition 12.1.7 (Weight). Let $X$ be a topological space and $\mathfrak{B}$ be the collection of all bases for the topology of $X$. We define the weight of $X$ as

$$
w(X)=\min _{\mathcal{B} \in \mathfrak{B}}|\mathcal{B}|
$$

Definition 12.1.8 (Basis at a point). Let $X$ be a topological space and $p \in X$ be any fixed point. We define the collection $\mathcal{B}_{p} \subseteq 2^{X}$ of neighbourhoods of $p$ to be the neighbourhood basis for the topology of $X$ at $p$ if for any neighbourhood $U_{p} \subseteq X$, there exists $B \in \mathcal{B}_{p}$ such that $B \subseteq U_{p}$.

Definition 12.1.9. Let $(X, \tau)$ be a topological space. Let $x \in X$ be any point and consider $\mathfrak{B}_{x}$ the collection of all bases at $x$. Then we define the character of $X$ at the point $x$ as

$$
\chi(x,(X, \tau))=\min _{\mathcal{B}_{x} \in \mathcal{B}_{x}}\left|\mathcal{B}_{x}\right|
$$

## Closed and Open Sets

The notion of a closed set and an open set are closely related - pardon for the pun. They have a dual relationship, allowing for us to define topologies via either of them. Lets first define what we mean by a closed set.

Definition 12.1.10 (Closed set). Let $X$ be a topological space. We define a set $A \subseteq X$ to be closed if $X \backslash A$ is open.

The following proposition realizes the idea that the duality of open and closed sets allow us to work with topological spaces by analysing both open and closed elements of the space of interest.

Proposition 12.1.11. If $X$ is a topological space, then

1. The sets $X$ and $\varnothing$ are closed.
2. The finite union of closed sets is closed.
3. The arbitrary intersection of closed sets is closed.

Proof. Notice that $X \backslash X=\varnothing$ and $X \backslash \varnothing=X$ are both open sets from Definition 12.1.1, hence $X$ and $\varnothing$ are closed. Let $\left\{C_{j}\right\}_{j=1}^{n}$ be a finite collection of closed sets, then $X \backslash \bigcup_{j=1}^{n} C_{j}=\bigcap_{j=1}^{n} X \backslash C_{j}$ but since $X \backslash C_{j}$ is open for all $j$, then their finite intersection is open and hence $X \backslash \bigcup_{j=1}^{n} C_{j}$ is also open, which implies by definition that $\bigcup_{j=1}^{n} C_{j}$ is closed. Consider now any collection of closed sets $\left\{C_{j}\right\}_{j \in J}$ - where $J$ is possibly infinite. Then $X \backslash \bigcap_{j \in J} C_{j}=\bigcup_{j \in J} X \backslash C_{j}$ is open by the arbitrary union of open sets being open, hence $\bigcap_{j \in J} C_{j}$ is closed.

We now define four important operations on sets of a topological space, which will accompany us for the rest of these notes on general point-set topology.

Definition 12.1.12. Let $X$ be a topological space and $A \subseteq X$ be a set. We define
(a) The closure of $A$ is the least closed set, $\mathrm{Cl} A$, that contains $A$. This can be equivalently described as

$$
\mathrm{Cl} A=\bigcap\{F \subseteq X: A \subseteq F \text { and } F \text { is closed }\} .
$$

(b) The interior of $A$ is the biggest open set $\operatorname{Int} A$ contained in $A$. That is

$$
\text { Int } A=\bigcup\{U \subseteq X: U \subseteq A, U \text { is open }\}
$$

(c) The exterior of $A$ in $X$ is defined as

$$
\operatorname{Ext} A=X \backslash \mathrm{Cl} A
$$

(d) The boundary of $A$ in $X$ is defined as

$$
\partial A=X \backslash(\operatorname{Int} A \cup \operatorname{Ext} A) .
$$

Remark 12.1.13. For the remainder of this section, unless specified on the contrary, we let $X$ be any topological space and $A \subseteq X$ be any subset of $X$.

Proposition 12.1.14. Int $A$ and $\operatorname{Ext} A$ are open sets, while on the other hand $\mathrm{Cl} A$ and $\partial A$ are closed sets.

Proof. Since $\operatorname{Int} A$ is the union of a collection of open sets, $\operatorname{Int} A$ is open. $\mathrm{Cl} A$ is the intersection of a collection of closed sets, hence $\mathrm{Cl} A$ is closed. The exterior set Ext $A$ is simply the complement of a closed set, hence it's open. The boundary of $A$ is the complement of the union of open sets (which is open), hence $\partial A$ is closed.

Proposition 12.1.15. The following are equivalences on the definitions of open and closed sets.
(a) Open set equivalences:

- $A$ is open.
- $A=\operatorname{Int} A$.
- $A$ contains none of its boundary points, i.e. $A \cap \partial A=\varnothing$.
- For all $x \in A$ there exists a neighbourhood $U \subseteq A$ containing $x$.
(b) Closed set equivalences:
- $A$ is closed.
- $A=\mathrm{Cl} A$.
- For all $x \in X \backslash A$, there exists a neighbourhood $U \subseteq X \backslash A$ of $x$.

Proof. (a) Let $A$ be open, then, in particular, $A \in\{U \subseteq X: U \subseteq A, U$ is open $\}$, thus $A=\operatorname{Int} A$. Moreover, from the definition of $\partial A$ it follows that $A \cap \partial A=$ $\operatorname{Int} A \cap A \backslash(\operatorname{Int} A \cup \operatorname{Ext} A)=\varnothing$. On the other hand, if $p \in A$ is any point, then $A$ itself is a neighbourhood of $p$.

In order to finish the equivalence chain, let $A$ be such that all of its points have a neighbourhood contained in $A$. We can then define the collection of neighbourhoods $\mathcal{U}=\left\{U_{p} \subseteq A: p \in A, p \in U\right\}$. Notice that $A \subseteq \bigcup_{U \in \mathcal{U}} U$ and the opposite inclusion is clearly true - thus $A$ is the union of a collection of open sets of $X$, hence $A$ is open.
(b) Let $A$ be closed, then $A$ is clearly the least closed set containing itself, implying in $A=\mathrm{Cl} A$. Let $x \in X \backslash A$ be any point. Since $X \backslash A$ is open, we use the previous equivalence for open sets to conclude that there exists $U \subseteq X \backslash A$ neighbourhood of $x$. To conclude the equivalence chain, suppose the last property is true for $X \backslash A$. Then $X \backslash A$ is open by the previous item, which, in turn, this implies that $A$ closed.

Proposition 12.1.16 (Basis criterion for open sets). Let $X$ a topological space and $\mathcal{B}$ a base for the topology of $X$. A set $A \subseteq X$ is open if and only if for all points $p \in A$ there exists a neighbourhood of $p, B_{p} \in \mathcal{B}$, such that $B_{p} \subseteq A$.

Proof. Let $A$ be open, then there exists a collection of elements of $\mathcal{B}$ whose union is $A$ - which implies that there exists, for all $p \in A$, a set $B \subseteq \mathcal{B}$ such that $p \in B$ and $B \subseteq A$. For the contrary, if $p \in A$ is any point and $B \in \mathcal{B}$ is the corresponding neighbourhood $p \in B \subseteq A$, then from Proposition 12.1.15 we find that $A$ is open.

Proposition 12.1.17. The following propositions are equivalent, regarding points on the closure of a set $A$
(a) $x \in \mathrm{Cl} A$.
(b) Every neighbourhood $U \subseteq X$ of $x$ is such that $U \cap A \neq \varnothing$.
(c) There exists a basis $\mathcal{B}_{x}$ at the point $x$ such that for all $U \in \mathcal{B}_{x}$ we have $U \cap A=\varnothing$.

Proof. (a) implies (b): Let $x \in X$ and suppose that (b) is false for $x$, so that there exists $U \subseteq X$ neighbourhood of $x$ such that $U \cap A=\varnothing$. Then $A \subseteq X \backslash U$, that is, $A$ is a subset of the complement of $U$. Since $U$ is open, then $X \backslash U \in C_{A}:=\{F \subseteq X: A \subseteq F, F$ is closed $\}$. From the definition of closure, we have that $\mathrm{Cl} A \subseteq X \backslash U$ and therefore $x \notin \mathrm{Cl} A$. (b) implies (c): From the definition of a basis at the point $x$, we know that every $U \in \mathcal{B}_{x}$ is a neighbourhood of $x$, hence if (b) is true for $x$, proposition (c) follows immediately. (c) implies (a): Let $x \in X$ such that $x \notin \mathrm{Cl} A$, so that proposition (a) is false for $x$. From the definition of closure, there exists $F \subseteq C_{A}$ such that $x \notin F$. Consider the open complement $V=X \backslash F$ so that $x \in V$ and $V \cap A=\varnothing$. Hence, given any basis at $x$ there exists a neighbourhood of $x$, say $U$, such that $U \subseteq V$ and hence $U \cap A=\varnothing$, which implies that proposition (c) is false for $x$.

Corollary 12.1.18. If $U$ is an open set and $U \cap A=\varnothing$, then $U \cap \mathrm{Cl} A=\varnothing$. Also, if $U$ and $V$ are disjoint sets, then $U \cap \mathrm{Cl} V=\mathrm{Cl} U \cap V=\varnothing$.

Proof. Suppose $U \cap A=\varnothing$ and that there exists $x \in U \cap \mathrm{Cl} A$, so that $x \in \mathrm{Cl} A$. Since $U$ is open, it is a neighbourhood of $x$, hence from Proposition 12.1.17 we find that $U \cap A \neq \varnothing$, which is false, thus $U \cap \mathrm{Cl} A=\varnothing$.

In order to classify points as being interior, exterior or on the boundary of a set, we may use the following important proposition.

Proposition 12.1.19 (Interior, exterior and boundary points). Classification of points:
(a) $x \in \operatorname{Int} A$ if and only if there exists $U \subseteq A$ neighbourhood of $x$.
(b) $x \in \operatorname{Ext} A$ if and only if there exists $U \subseteq X \backslash A$ neighbourhood of $x$.
(c) $x \in \partial A$ if and only if all neighbourhoods $U \subseteq X$ of $x$ are such that $U \cap A \neq \varnothing$ and $U \cap(X \backslash A) \neq \varnothing$.

Proof. (a) Let $x \in A$ be any point. Suppose $U \subseteq X$ is neighbourhood of $x$ such that $U \subseteq A$, then from the fact that $U$ is open we conclude that $U \subseteq \operatorname{Int} A$ from the definition of the interior operator. Suppose $x \in \operatorname{Int} A$ then from definition there is a neighbourhood of $x$ contained in $A$.
(b) Suppose exists $U \subseteq X \backslash A$ neighbourhood of $x$. From Proposition 12.1.17 we find $x \notin \mathrm{Cl} A$ and therefore $x \in \operatorname{Ext} A$. The other side of the implication is trivial.
(c) Since all neighbourhoods of $x$ contain a point of $X \backslash A$, from the first item of this proposition we find that $x$ cannot belong to the interior of $A$. Moreover, every neighbourhood contains a point of $A$, then $x$ cannot be an element of $\operatorname{Ext} A$. Therefore $x \in \partial A$.

Proposition 12.1.20 (Decomposition of the closure). $\mathrm{Cl} A=A \cup \partial A=\operatorname{Int} A \cup \partial A$.
Proof. For the first equality we have

$$
\begin{aligned}
A \cup \partial A & =A \cup[X \backslash(\operatorname{Int} A \cup \operatorname{Ext} A)] \\
& =A \cup[(X \backslash \operatorname{Int} A) \cap(X \backslash \operatorname{Ext} A)] \\
& =A \cup[(X \backslash \operatorname{Int} A) \cap \mathrm{Cl} A] \\
& =[A \cup(X \backslash \operatorname{Int} A)] \cap[A \cup \mathrm{Cl} A] \\
& =X \cap \mathrm{Cl} A=\mathrm{Cl} A .
\end{aligned}
$$

Analogously, for the second equality

$$
\begin{aligned}
\operatorname{Int} A \cup \partial A & =\operatorname{Int} A \cup[(X \backslash \operatorname{Int} A) \cap \mathrm{Cl} A] \\
& =[\operatorname{Int} A \cup(X \backslash \operatorname{Int} A)] \cap[\operatorname{Int} A \cup \mathrm{Cl} A] \\
& =X \cap \mathrm{Cl} A=\mathrm{Cl} A .
\end{aligned}
$$

This proves the proposition.
Proposition 12.1.21. Let $X$ be a topological space and $A \subseteq X$. Then $\mathrm{Cl}(X \backslash A)=X \backslash \operatorname{Int} A$ and also $\operatorname{Int}(X \backslash A)=X \backslash \mathrm{Cl} A$.

Proof. We prove the first equality. Let $\mathcal{A}:=\{U \subseteq A: U$ is open $\}$, from definition we have that $\operatorname{Int} A=\bigcup_{U \in \mathcal{A}} U$, moreover, $X \backslash \operatorname{Int} A=X \backslash \bigcup_{U \in \mathcal{A}} U=\bigcap_{U \in \mathcal{A}} X \backslash U$ from de Morgan's Laws. Notice that obviously $X \backslash U \subseteq X$ and moreover since $U$ is open, then the complement $X \backslash U$ is closed. This makes $\underset{\sim}{X} \backslash \operatorname{Int}(A)=\bigcap_{U \in \mathcal{A}} X \backslash U=$ $\mathrm{Cl}(X \backslash A)$. Now we show the second equality. Define $\tilde{\mathcal{A}}:=\{U \subseteq A: U$ is closed $\}$, then $\mathrm{Cl} A=\bigcap_{U \in \tilde{\mathcal{A}}} U$ and hence $X \backslash \mathrm{Cl} A=X \backslash \bigcap_{U \in \tilde{\mathcal{A}}} U=\bigcup_{U \in \tilde{\mathcal{A}}} X \backslash U$. Notice that $X \backslash U \subseteq X \backslash A$ and since $U$ is closed, then $X \backslash U$ is open. From this we conclude that $X \backslash \mathrm{Cl} A=\bigcup_{U \in \tilde{\mathcal{A}}} X \backslash U=\operatorname{Int}(X \backslash A)$.

Proposition 12.1.22. Let a finite collection of subsets $\left\{A_{i}\right\}_{i=1}^{n} \subseteq 2^{X}$, where $X$ is a topological space, then we have that

$$
\mathrm{Cl}\left(\bigcup_{i=0}^{n} A_{i}\right)=\bigcup_{i=0}^{n} \mathrm{Cl} A_{i}
$$

Proof. It suffices to prove that for $A, B \subseteq X$ we have $\mathrm{Cl}(A \cup B)=\mathrm{Cl} A \cup \mathrm{Cl} B$. Notice that $\mathrm{Cl} A, \mathrm{Cl} B \subseteq \mathrm{Cl}(A \cup B)$, hence $\mathrm{Cl} A \cup \mathrm{Cl} B \subseteq \mathrm{Cl}(A \cup B)$. Now, since $A \subseteq \mathrm{Cl} A$ and $B \subseteq \mathrm{Cl} B$, we find $A \cup B \subseteq \mathrm{Cl} A \cup \mathrm{Cl} B$, then $\mathrm{Cl}(A \cup B) \subseteq \mathrm{Cl}(\mathrm{Cl} A \cup \mathrm{Cl} B)=\mathrm{Cl} A \cup \mathrm{Cl} B$.

Definition 12.1.23 (Locally finite family). A collection of subsets $\left\{A_{i}\right\}_{i \in I} \subseteq 2^{X}$ of a topological space $X$ is said to be a locally finite family if for every point $x$ there exists a neighbourhood $U \subseteq X$ for which the collection $\left\{i \in I: U \cap A_{i} \neq \varnothing\right\}$ is finite.

Definition 12.1.24 (Discrete family). A collection of subsets $\left\{A_{i}\right\}_{i \in I} \subseteq 2^{X}$ of a topological space $X$ is said to be a discrete family if for all $x \in X$ there exists a neighbourhood $U \subseteq X$ such that there exists at most one $A_{i}$ in the family such that $U \cap A \neq \varnothing$. A discrete family is also a locally finite family.

Proposition 12.1.25. Given a locally finite family of sets $\left\{A_{i}\right\}_{i \in I} \subseteq 2^{X}$, where $X$ is a topological space, we have that

$$
\mathrm{Cl}\left(\bigcup_{i \in I} A_{i}\right)=\bigcup_{i \in I} \mathrm{Cl} A_{i}
$$

Proof. Notice that clearly $A_{i} \subseteq \mathrm{Cl}\left(\bigcup_{i \in I} A_{i}\right)$ for all $i \in I$, hence $\bigcup_{i \in I} \mathrm{Cl} A_{i} \subseteq \mathrm{Cl}\left(\bigcup_{i \in I} A_{i}\right)$. Moreover, we have from hypothesis that $\left\{A_{i}\right\}_{i \in I}$ is a locally finite family, thus given $x \in \mathrm{Cl}\left(\bigcup_{i \in I} A_{i}\right)$ we can find a neighbourhood of $x$, say $U \subset X$, such that $I_{0}:=\{i \in I$ : $\left.U \cap A_{i} \neq \varnothing\right\}$ is finite. Notice that from Proposition 12.1.17 we have that $x$ cannot be a limit point of any of the sets with index $i \in I \backslash I_{0}$, hence $x \notin \mathrm{Cl}\left(\cup_{i \in I \backslash I_{0}} A_{i}\right)$. On the other hand, we have $x \in \mathrm{Cl}\left(\bigcup_{i \in I} A_{i}\right)=\mathrm{Cl}\left(\bigcup_{i \in I_{0}} A_{i}\right) \cup \mathrm{Cl}\left(\bigcup_{i \in I \backslash I_{0}} A_{i}\right)$, which implies that $x \in \mathrm{Cl}\left(\bigcup_{i \in I_{0}} A_{i}\right)=\bigcup_{i \in I_{0}} \mathrm{Cl} A_{i} \subseteq \bigcup_{i \in I} \mathrm{Cl} A_{i}$ (the equality comes from the fact that $I_{0}$ is finite and hence Proposition 12.1.22 hold).

Proposition 12.1.26. If $\left\{A_{i}\right\}_{i \in I}$ is a locally finite (resp. discrete) family, then the family $\left\{\mathrm{Cl} A_{i}\right\}_{i \in I}$ is locally finite (resp. discrete). Conversely, if $\left\{\mathrm{Cl} A_{i}\right\}_{i \in I}$ is locally finite, then $\left\{A_{i}\right\}_{i \in I}$ is locally finite.

Proof. Let the locally finite (resp. discrete) family $\left\{A_{i}\right\}_{i \in I}$. Given any element $x \in X$ and a neighbourhood $U$ of $x$ such that the indexing set $I_{0}:=\left\{i \in I: U \cap A_{i} \neq \varnothing\right\}$ is finite (resp. is either empty or a singleton). Notice that since $U \cap A_{i} \subseteq U \cap \mathrm{Cl} A_{i}$, we find that for all $i \in I_{0}, U \cap \mathrm{Cl} A_{i} \neq \varnothing$, so that $I_{0}=I_{0}^{\prime}:=\left\{i \in I: U \cap \mathrm{Cl} A_{i} \neq \varnothing\right\}$. Since $U$ is an open set, from Corollary 12.1.18 we find that $U \cap \mathrm{Cl} A_{i}=U \cap A_{i}=\varnothing$ for all $i \in I \backslash I_{0}$. For the converse, if the collection of closures is locally finite, then clearly $\left\{A_{i}\right\}_{i \in I}$ is locally finite.

## Derived and Dense Sets

Definition 12.1.27 (Limit point and derived set). A point $p \in A$ is called a limit point of $A$ if every neighbourhood of $p$ has a point of $A \backslash\{p\}$. We define the set $A^{\prime}$ as the collection of all limit points of $A$, which we'll call derived set.

Proposition 12.1.28. A point $p \in A$ is a limit point if and only if $p \in \operatorname{Cl}(A \backslash\{p\})$.
Proof. If $p$ is a limit point, then from Proposition 12.1 .17 we see that $p \in \operatorname{Cl}(A \backslash\{p\})$. Moreover, if $p \in \mathrm{Cl}(A \backslash\{p\})$, then again from Proposition 12.1.17 we see that there is a basis $\mathcal{B}_{p}$ at $p$ such that for all $U \in \mathcal{B}_{p}$ we have $U \cap(A \backslash\{p\}) \neq \varnothing$ - thus $p$ is a limit point of $A$.

Definition 12.1.29 (Isolated points). If $p \in A \backslash A^{\prime}$, we say that $p$ is an isolated point of A.

Proposition 12.1.30. A point $p \in A$ is isolated if and only if there exists a neighbourhood of $p$ for which the only point of intersection with $A$ is $p$, i.e. $U \subseteq X$ neighbourhood of $p$ with $U \cap A=\{p\}$.

Proof. If $p$ is isolated then there exists a neighbourhood $U \subseteq X$ of $p$ such that $U \cap(A \backslash$ $\{p\})=\varnothing$ but since $p \in U$, then it follows that $U \cap A=\{p\}$. Now, suppose there exists such $U$, then in particular $p$ does not satisfy the condition to be a limit point of $A$, thus $p \notin A$, hence $p \in A \backslash A^{\prime}$ and therefore $p$ is a limit point.

Proposition 12.1.31. The closure of a set is the union of the set with its limit points, i.e. $\mathrm{Cl} A=A \cup A^{\prime}$.

Proof. Let $p \in A^{\prime}$, then from definition any neighbourhood of $p$ intersects $A \backslash\{p\}-$ and, in particular, intersects $A$ - therefore, by means of Proposition 12.1.17 we see that $p \in \mathrm{Cl} A$, thus $A^{\prime} \subseteq \mathrm{Cl} A$. It's obvious that $A \subseteq \mathrm{Cl} A$, thus $A \cup A^{\prime} \subseteq \mathrm{Cl} A$. On the other hand, if $q \in \mathrm{Cl} A$, then we assume that $q \notin A$ - since, on the contrary, it's clear that $q \in A \cup A^{\prime}$. Again using Proposition 12.1.17 we get that every neighbourhood of $q$ intersects $A$, and since $q \notin A$, then the intersection contains a point other than $q$ that is, $q$ is a limit point of $A$. Hence $\mathrm{Cl} A \subseteq A \cup A^{\prime}$, which finishes the proof.

Corollary 12.1.32. A set $A$ is closed if and only if $A$ contains all of its limit points, that is, $A=A^{\prime}$.

Proof. If $A$ is closed, then $A=\mathrm{Cl} A$ and hence $A^{\prime} \subseteq A$ from Proposition 12.1.31. Otherwise, if $A$ contains all of its limit points, then $A \cup A^{\prime}=A=\mathrm{Cl} A$ - that is, $A$ is closed.

Definition 12.1.33 (Dense set). A set $A \subseteq X$ is said to be dense if $\mathrm{Cl} A=X$.
Proposition 12.1.34. A set $A$ is dense in $X$ if and only if for all non-empty open subsets of $X$ contains a point of $A$.

Proof. Let $A$ be dense in $X$ and let $U \subseteq X$ be any non-empty open subset of the space. Suppose, for the sake of contradiction, that $U \cap A=\varnothing$. Since $U$ is non-empty, take any $x \in U$, then from Proposition 12.1 .17 we find that $x \notin \mathrm{Cl} A$ but since $\mathrm{Cl} A=X$, then $x \notin X$, which is a contradiction - thus $U \cap A$ is non-empty. On the other hand, if we suppose that every non-empty open set of the space intersects $A$, then given any $x \in X$ we see that $x$ is a limit point of $A$, hence $x \in \mathrm{Cl} A$, thus $X \subseteq \mathrm{Cl} A$, which proves the proposition since $\mathrm{Cl} A \subseteq X$.

### 12.2 Morphisms of Topological Spaces

Definition 12.2.1 (Continuous map). Let $X$ and $Y$ be topological spaces and consider the map $f: X \rightarrow Y$. We say that $f$ is continuous if for all $U \subseteq Y$ open, the preimage $f^{-1}(U)$ is open in $X$.

Proposition 12.2.2. A map $f: X \rightarrow Y$ is continuous if and only if the preimage of every closed subset is closed.

Proof. Suppose that $f$ satisfies the latter, the given any $U \subseteq Y$ there exists a closed set $C \subseteq Y$ such that $U=Y \backslash C$, thus $f^{-1}(C)$ is closed, then $f^{-1}(X \backslash C)=f^{-1}(U)$ is open. For the contrary, the analogous argument is used.

Proposition 12.2.3. Let $X, Y, Z$ be topological spaces. The following properties of continuous maps between topological spaces hold
(CM1) Every constant map $f: X \rightarrow Y$ is continuous.
(CM2) The identity map is continuous.
(CM3) If $f: X \rightarrow Y$ is continuous, then for all open set $U \subseteq X$ we have $\left.f\right|_{U}$ continuous.
(CM4) If $f: X \rightarrow Y$ and $g: Y \rightarrow Z$ are continuous maps, then $g f: X \rightarrow Z$.
Proof. (1) Let any point $a \in Y$ and consider the constant map $x \stackrel{f}{\mapsto} a$ for all $x \in X$. Then, for all $U \subseteq Y \backslash\{a\}$ open, we have that $f^{-1}(U)=\varnothing$, thus have open preimage. On the other hand, the fibre $f^{-1}(a)=X$, hence also open, thus $f$ is continuous.
(2) Notice that if $U \subseteq Y$ is any open set, then $\mathrm{id}_{X}^{-1}(U)=U$ and thus open.
(3) Let $g: X \rightarrow Y$ be a continuous map, and $U \subseteq X$ be any open set, then we can take any open $V \subseteq g(U)$ and conclude that $g^{-1}(V)$ is open (from the hypothesis that $g$ is continuous); let now an open $H \subseteq Y \backslash g(U)$, then certainly $\left.g\right|_{U} ^{-1}(H)=\varnothing$, thus open. Hence $g \mid u: U \rightarrow Y$ is indeed continuous.
(4) Let $U \subseteq Z$ be open, then $(g f)^{-1}(U)=f^{-1}\left(g^{-1}(U)\right)$. Moreover, since from hypothesis $g$ is continuous, then $g^{-1}(U)$ is open, now, from the continuity of $f$ we conclude that $f^{-1}\left(g^{-1}(U)\right)$ is open, hence $g f$ is continuous.

Definition 12.2.4 (Category of topological spaces). We define the category of topological spaces to be composed of objects named topological spaces and morphisms being continuous maps between them - we'll denote such category by Top.

Proposition 12.2.5 (Local criterion for continuity). Let $f: X \rightarrow Y$ be a map between topological spaces. Then $f$ is continuous if and only if for all points $x \in X$ there exists a neighbourhood $U_{x} \subseteq X$ of $x$ such that $\left.f\right|_{U_{x}}$ is continuous.

Proof. If $f$ is a morphism, then for any $x \in X$ we can take the whole space $X$ as a neighbourhood of $x$ and the proposition follows. On the other hand If $f$ is locally continuous for every point of $X$, let $V \subseteq Y$ be any open set and consider, for every $x \in f^{-1}(V)$ a neighbourhood $U_{x}$ such that $\left.f\right|_{U_{x}}$ is continuous - and hence $\left.f\right|_{U_{x}} ^{-1}(V)$ is open for all $x \in f^{-1}(V)$. Notice that $f^{-1}(V) \cap U_{x}=\left.f\right|_{U_{x}} ^{-1}(V)$, which has to be open on $X$. Moreover, from construction $f^{-1}(V)=\bigcup_{x \in f^{-1}(V)} f^{-1}(V) \cap U_{x}$ is the union of open sets, thus $f^{-1}(V)$ is open - and therefore $f$ is continuous.

Proposition 12.2.6 (Basis criterion for continuity). Let $f: X \rightarrow Y$ be a map between topological spaces, and $\mathcal{B}$ be a basis for the topology of $Y$. Then, $f$ is continuous (hence a morphism) if and only if for all $B \in \mathcal{B}$ we have $f^{-1}(B) \subseteq X$ open.

Proof. $(\Rightarrow)$ If $f$ is continuous, then certainly $f^{-1}(B)$ is open. $(\Leftarrow)$ On the other hand, if $A \subseteq X$ is any open set, then $A=\bigcup_{p \in A} B_{p}$ for $B_{p} \in \mathcal{B}$ neighbourhood of $p$. Hence $f^{-1}(A)=f^{-1}\left(\bigcup_{p \in A} B_{p}\right)=\bigcup_{p \in A} f^{-1}\left(B_{p}\right)$ is the union of open sets, thus $f^{-1}(A)$ is open and therefore $f$ is continuous.

Proposition 12.2.7. Let $f: X \rightarrow Y$ be a morphism of topological spaces and $\mathcal{B}$ be a basis for the space $X$. Then $f(\mathcal{B})=\{f(B): B \in \mathcal{B}\}$ is a basis for the space $Y$ if and only if $f$ is surjective and open.

Proof. $(\Rightarrow)$ Suppose $f(\mathcal{B})$ is a basis for $Y$, then for all $V \subseteq Y$ open, there exists an indexing set $I$ such that $V=\bigcup_{i \in I} U_{i}$, where $U_{i} \in f(\mathcal{B})$. This implies in the existence of $B \in \mathcal{B}$ such that $U_{i}=f(B)$ hence $f$ is open. Consider now any point $y \in Y$ and any neighbourhood $V_{y}$ of $y$. From the same argument as above we have $V_{y}=\bigcup_{i \in I_{y}} U_{i}$, where there exists some $i \in I_{y}$ such that $y \in U_{i}$ and hence $y \in U_{i}=f(B)$ for some $B \in \mathcal{B}$. $(\Leftarrow)$ Suppose $f$ is surjective and open. Let $V \subseteq Y$ be any open set. Since $f$ is continuous and surjective, we have that $U=f^{-1}(V)$ is a non-empty open set. Since $\mathcal{B}$ is a basis, we can write $U=\bigcup_{i \in I} B_{i}$ where $B_{i} \in B$, and then $f(U)=f\left(\bigcup_{i \in I} B_{i}\right)=\bigcup_{i \in I} f\left(B_{i}\right)$, where $f\left(B_{i}\right) \in f(\mathcal{B})$ and $f(U)=V$ from the fact that $f$ is surjective. Since $f$ is open, $f(B)$ is open for all $B \in \mathcal{B}$. Hence $f(\mathcal{B})$ is a basis for the space $Y$.

Definition 12.2.8 (Isomorphism). Let $f$ be a morphism of topological spaces. If $f$ is bijective and has a continuous inverse, then we say that $f$ is an isomorphism of topological spaces - which can also be called an homeomorphism.

Definition 12.2.9 (Open \& Closed maps). Let $f: X \rightarrow Y$ be any set-function between topological spaces.

- We say that $f$ is an open map if for all $U \subseteq X$ open the image $f(U) \subseteq Y$ is open.
- We say that $f$ is a closed map if for all $C \subseteq X$ closed, the image $f(C) \subseteq Y$ is closed.

Proposition 12.2.10. Let $f: X \rightarrow Y$ be a map of topological spaces and consider that $f$ is an isomorphism, then $f$ is an open and a closed map.

Proof. Let $U \subseteq X$ be an open (resp. closed) set and consider $V:=f(U) \subseteq Y$. Since $f$ is a bijection, we find that $f(U)=V$ is open (resp. closed) by the continuity of $f^{-1}$. Hence $f$ is an open (resp. closed) map.

Corollary 12.2.11. Let $f: X \rightarrow Y$ be a bijective set-function, where $X$ and $Y$ are topological spaces. The following propositions are equivalent
(a) $f$ is an isomorphism of topological spaces.
(b) $f$ is open.
(c) $f$ is closed.

Proof. (a) $\Rightarrow$ (b): Suppose $f$ is an isomorphism, then since $f^{-1}$ is also an isomorphism, it follows that, given any $U \subseteq X$ the image $f(U)$ is open - hence $f$ is open. (b) $\Rightarrow$ (c): Let $f$ be open, then given any closed set $C \subseteq X$, we have that $f(X \backslash C) \subseteq Y$ is open, moreover, since $f$ is injective, $f(X \backslash C)=f(X) \backslash f(C)$ and since $f(X)=Y$ from the surjectivity of $f$, we conclude that $f(C)$ is closed. (c) $\Rightarrow$ (a) Let $f$ be closed, then given any $V \subseteq Y$ open, we know that $Y \backslash V$ is closed and since $f$ is a bijection, $f^{-1}(Y \backslash V)=f^{-1}(Y) \backslash f^{-1}(V)=X \backslash f^{-1}(V) \subseteq X$ is closed, thus $f^{-1}(V)$ is open - which implies in the continuity of $f$. Moreover, since $f$ is a bijection, $f^{-1}$ is also closed, and therefore continuous by the same analogous proof, thus $f$ is an isomorphism.

Proposition 12.2.12. Let $X$ and $Y$ be topological spaces and $f: X \rightarrow Y$ be a set-function. We can classify the behaviour of $f$ by the following conditions
(a) $f$ is a morphism of topological spaces if and only if $f(\mathrm{Cl} A) \subseteq \mathrm{Cl}(f(A))$ for every set $A \subseteq X$.
(b) $f$ is a morphism of topological spaces if and only if $f^{-1}(\operatorname{Int} B) \subseteq \operatorname{Int} f^{-1}(B)$ for every set $B \subseteq Y$.
(c) $f$ is closed if and only if $f(\mathrm{Cl} A) \supseteq \mathrm{Cl}(f(A))$ for every set $A \subseteq X$.
(d) $f$ is open if and only if $f^{-1}(\operatorname{Int} B) \supseteq \operatorname{Int} f^{-1}(B)$ for every set $B \subseteq Y$.

Proof.

## prove

Definition 12.2.13 (Local isomorphism). Let $f: X \rightarrow Y$ be a set-function between topological spaces $X$ and $Y$. We say that $f$ is a local isomorphism of topological spaces if, for all $x \in X$, there exists a neighbourhood $U \subseteq X$ such that $f(U) \subseteq Y$ is open and the induced map $f: U \xrightarrow{\simeq} f(U)$ is an isomorphism of topological spaces.

Proposition 12.2.14. The following are properties pertaining to local isomorphisms
(a) Every isomorphism is a local isomorphism.
(b) Every local isomorphism is continuous and open.
(c) Every bijective local isomorphism is an isomorphism.

Proof. (a) Let $f: X \xrightarrow{\simeq} Y$ be an isomorphism, then for all $x \in X$ we can choose the neighbourhood $X$ and the restriction $\left.f\right|_{X}=f: X \xrightarrow{\cong} f(X)=Y$ is an isomorphism.
(b) Let $g: X \rightarrow Y$ be a local isomorphism. Let $x \in X$ be any element and consider $U \subseteq$ $X$ such that $\left.g\right|_{U}: U \xrightarrow{\simeq} g(U)$ is an isomorphism, then in particular $\left.g\right|_{U}$ is continuous - by Proposition 12.2 .5 we find that $g$ is continuous. Now let $U \subseteq X$ be any open set, for each $x \in U$ take $U_{x} \subseteq X$ neighbourhood of $x$ such that $\left.g\right|_{U_{x}}: U_{x} \rightarrow g\left(U_{x}\right)$ is an isomorphism. Notice that the restriction $\left.g\right|_{U \cap U_{x}}: U \cap U_{x} \xrightarrow{\simeq} f\left(U \cap U_{x}\right)$ is also an isomorphism which implies in $f\left(U \cap U_{x}\right) \subseteq Y$ being open - moreover, $V=\bigcup_{x \in U} f\left(U \cap U_{x}\right)$ thus $V$ is open.
(c) Let $f: X \rightarrow V$ be a bijective local isomorphism, then by the last item, $f$ is open. Using Corollary 12.2.11 we see that $f$ is an isomorphism.

Definition 12.2.15 (Embedding). Let $f: X \hookrightarrow Y$ be an injective morphism of topological spaces. If the induced morphism $f: X \rightarrow f(X)$ is an isomorphism, then we say that $f$ is a topological embedding of $X$ in $Y$.

## Topology Generated by Mappings

Proposition 12.2.16 (Topology generated by a collection of mappings). Let $X$ be a topological space and $\left\{Y_{i}\right\}_{i \in I}$ be a collection of topological spaces. Let $\left\{f_{i}: X \rightarrow Y_{i}\right\}_{i \in I}$ be the collection of mappings between such topological spaces. Then there exists a initial topology on $X$ such that $f_{i}$ is continuous, for all $i \in I$. Such topology is generated by the base

$$
\mathcal{B}=\left\{\bigcap_{j \in J} f_{j}^{-1}\left(U_{j}\right): U_{i} \subseteq Y_{i} \text { is open, } J \subseteq I \text { is finite }\right\}
$$

We call such topology as the topology generated by the collection of mappings $\left\{f_{i}\right\}_{i \in I}$.
Proof. First we show that $\mathcal{B}$ is indeed a basis for $X$. Let $x \in X$ be any point, then clearly $x \in f_{i}^{-1}\left(Y_{i}\right)$ for all $i \in I$, hence $x \in \bigcap_{j \in J} f_{j}^{-1}\left(Y_{j}\right) \in \mathcal{B}$ for a finite indexing set $J \subseteq I$. Let $J, S \subseteq I$ be finite indexing sets, then consider the sets $A:=\bigcap_{j \in J} f_{j}^{-1}\left(U_{j}\right), B:=$ $\bigcap_{s \in S} f_{s}^{-1}\left(V_{s}\right) \in \mathcal{B}$ and let any point $x \in A \cap B$. Then in particular we can let an non-empty indexing set $T=J \cap S$ so that $x \in f_{t}^{-1}\left(U_{t}\right)$ for all $t \in T$ and therefore $C:=\bigcap_{t \in T} f_{t}^{-1}\left(U_{t}\right)$ is such that $x \in C \subseteq A \cap B$. This shows that $\mathcal{B}$ indeed satisfies the basis properties.

Now we show the initial topology property. Let $\tau$ be the topology generated by $\mathcal{B}$ and consider $\tau^{\prime}$ to be any other topology on $X$ for which the functions $f_{i}$ are continuous
for all $i \in I$. Trivially we must have $\mathcal{B} \subseteq \tau^{\prime}$ and therefore $\tau \subseteq \tau^{\prime}$. This says that $\tau$ is coarser than $\tau^{\prime}$, which proves the proposition.

Proposition 12.2.17. Let $X$ and $Y$ be topological spaces, and the topology of $Y$ be generated by the collection of maps $\left\{f_{i}: Y \rightarrow Y_{i}\right\}_{i \in I}$, where $\left\{Y_{i}\right\}_{i \in I}$ is a collection of topological spaces. Then a map $f: X \rightarrow Y$ is continuous if and only if the composition $f_{i} f: X \rightarrow Y_{i}$ is continuous for every $i \in I$.

Proof. $(\Leftarrow)$ Suppose $f_{i} f$ is continuous for all $i \in I$, then since $f_{i}$ is continuous on the topology of $Y$ it follows that for any given open set $U \subseteq Y_{i}$ we have $f_{i}^{-1}(U)=V \subseteq Y$ open. In particular, notice that $\left(f_{i} f\right)^{-1}(U)=f^{-1}\left(f_{i}^{-1}(U)\right)=f^{-1}(V) \subseteq X$ which must be open from the hypothesis of the continuity of $f_{i} f .(\Rightarrow)$ Moreover, if $f$ is continuous, then clearly the composition of continuous functions is continuous.

### 12.3 Metric Spaces

## Write on metric spaces

Definition 12.3.1. Let $M$ be a set. We say that a map $d$ : $M \times M \rightarrow \mathbf{R}_{\geqslant 0}$ is a pre-metric if for all points $x, y, z \in M$ it satisfies
(a) Symmetry: $d(x, y)=d(y, x)$.
(b) Triangle inequality: $d(x, z) \leqslant d(x, y)+d(y, z)$.

Moreover, if $d$ happens to satisfy the condition that
(c) $d(x, y)=0$ implies in $x=y$.
then we say that $d$ establishes a metric in $M$.
The set $M$ together with the (pre)metric $d$ is called a (pre)metric space.

### 12.4 Hausdorff Spaces

Definition 12.4.1. Let $X$ be a topological space. We define
(1) $X$ is $\mathrm{T}_{0}$ (or Kolmogorov) if and only if for every pair of distinct points $x, y \in X$ there exists an open set containing one, but not both, of them.
(2) $X$ is $\mathrm{T}_{1}$ (or Fréchet) if and only if for every pair of distinct points $x, y \in X$ there exists open sets $U, V \subseteq X$ such that $x \in U$ and $y \in V$, but $x \notin V$ and $y \notin U$

Proposition 12.4.2. A space $X$ is $\mathrm{T}_{1}$ if and only if the constant sequence $(x)_{i \in \mathbf{N}}$ converges to $x$ and only to $x$.

Proof. $(\Rightarrow)$ Suppose $X$ is $\mathrm{T}_{1}$, then certainly $x \rightarrow x$. Let $y \in X$ be distinct of $x$, then by the $\mathrm{T}_{1}$ property, there exists a neighbourhood $U \ni y$ such that $x \notin U$, hence $(x)_{i \in \mathbf{N}}$ does not converge to $y .(\Leftarrow)$ Suppose $X$ is not $T_{1}$, then choose $y \in X$ for which every neighbourhood $U \ni y$ contains $x$, then $x \rightarrow y$.

Definition 12.4.3 (Hausdorff space). We say that a topological space $X$ is Hausdorff (or $T_{2}$ ) if for all pair of distinct points $x, y \in X$, there exists neighbourhoods $U_{x}$ and $U_{y}$, subsets of $X$, such that $U_{x} \cap U_{y}=\varnothing$.

Corollary 12.4.4. Every open subset of a Hausdorff space is Hausdorff.
Proof. Let $X$ be Hausdorff and $U \subseteq X$ be an open. Choose any distinct pair $x, y \in U$ and let $U_{x}, U_{y} \subseteq X$ be neighbourhoods of $x, y$ such that $U_{x} \cap U_{y}=\varnothing$. Since $U_{x} \cap U, U_{y} \cap U \subseteq$ $U$ are neighbourhoods of $x$ and $y$, respectively, then $\left(U_{x} \cap U\right) \cap\left(U_{y} \cap U\right)=\varnothing$, hence $U$ is Hausdorff.

Proposition 12.4.5 (Hausdorff properties). Let $X$ be a Hausdorff space. Then
(a) Every finite subset of $X$ is closed.
(b) If a sequence $\left(x_{i}\right) \subseteq X$ converges to $x \in X$, then the limit is unique.

Proof. (a) Let $p_{0} \in X$ be any point and $p \in X$ be a distinct point. Since $X$ is Hausdorff, choose $U_{p}, U_{p_{0}} \subseteq X$ such that $U_{p} \cap U_{p_{0}}=\varnothing$, hence $U_{p} \subseteq X \backslash\left\{p_{0}\right\}$. Since for all $x \in X \backslash\left\{p_{0}\right\}$ there exists a neighbourhood $U_{x} \subseteq X \backslash\{p\}$, hence $\{p\}$ must be a closed set (every point of $X \backslash A$ has a neighbourhood in $X \backslash A$ is equivalent of saying that $A$ is closed).
(b) Suppose that $x_{i} \rightarrow x$ and $x_{i} \rightarrow y$ are two limits of the sequence. Since $X$ is Hausdorff, if $x \neq y$ implies that there exists neighbourhoods $U_{x}, U_{y} \subseteq X$ for which $U_{x} \cap U_{y}=\varnothing$. From the definition of convergence, there exists $N \in \mathbf{N}$ such that for all $i \geqslant N, x_{i} \in U_{x}$ and exists $M \in \mathbf{N}$ for which every $i \geqslant M$, we have $x_{i} \in U_{y}$. Then, for $i \geqslant \max (N, M), x_{i} \in U_{x} \cap U_{y}$, which cannot happen if $x \neq y$, hence $x=y$.

Proposition 12.4.6 (Neighbourhoods of limit points). Let $X$ be a Hausdorff space and $A \subseteq X$. If $p \in X$ is a limit point of $A$, then every neighbourhood of $p$ contains infinitely many points of $A$.

Proof. Let $p$ be a limit point of $A$, then for any neighbourhood $U_{p} \subseteq X$ we must have $\left(U_{p} \backslash\{p\}\right) \cap A \neq \varnothing$. For the sake of contradiction, suppose that there exists a finite number of points, $n>1$, in the set $U_{p} \cap A=\left\{x_{i}\right\}_{i=1}^{n}$, which implies that $U_{p} \cap A$ is closed. Consider now the non-empty closed set $\left\{x_{i}\right\}_{i=1}^{n} \backslash\{p\}$, then $X \backslash\left(\left\{x_{i}\right\}_{i=1}^{n} \backslash\{p\}\right)$ is open and contains $p$, thus is a neighbourhood of $p$. Notice that since the intersection of finitely many open sets is open, then $U_{p} \cap\left(X \backslash\left(\left\{x_{i}\right\}_{i=1}^{n} \backslash\{p\}\right)\right)$ is also open and is a neighbourhood of $p$. However, notice that

$$
A \cap U_{p} \cap\left(X \backslash\left(\left\{x_{i}\right\}_{i=1}^{n} \backslash\{p\}\right)\right)=\left\{x_{i}\right\}_{i=1}^{n} \cap\left(\left(X \backslash\left\{x_{i}\right\}_{i=1}^{n}\right) \cup\{p\}\right)=\{p\}
$$

which is a contradiction to the fact that $p$ is a limit point of $A$ (because any neighbourhood of $p$ should also contain a point of $A$ other than $p$ ). Hence we conclude that $A \cap U_{p}$ must be infinite.

Proposition 12.4.7. Let $X$ be a Hausdorff space and $A \subseteq X$. Then the set of limit points of $A$, denoted by $A^{\prime}$, is closed in $X$.

Proof. Consider the complement set $X \backslash A^{\prime}$, we must show that it is open. Consider $x \in$ $X \backslash A^{\prime}$, so that $x$ is not a limit point of $A$ and hence there exists $U \subseteq X$ neighbourhood of $x$ such that $(U \backslash\{x\}) \cap A=\varnothing$. We now show that $U \subseteq X \backslash A^{\prime}$. Let $p \in U$ be any point, then the set $U \backslash\{x\}$ is a neighbourhood of $p$ and is disjoint with $A$, hence $p$ is not a limit point of $A$, that is, $p \notin A^{\prime}$, which proves that $U \subseteq X \backslash A^{\prime}$. Moreover, since $X$ is Hausdorff, the singleton $\{x\}$ is closed, hence $U \backslash\{x\}$ is open

Proposition 12.4.8. Let $f, g: X \rightarrow Y$ be morphisms of topological spaces, and $Y$ be a Hausdorff space. Then the set $\{x \in X: f(x)=g(x)\}$ is closed in $X$.

Proof. Consider the complement set $A:=X \backslash\{x \in X: f(x)=g(x)\}$. Let any point $x \in A$, since $f(x) \neq g(x)$ and $Y$ is Hausdorff, there exists neighbourhoods $U_{1}, U_{2} \subseteq$ $Y$ of $f(x), g(x)$, respectively, such that $U_{1} \cap U_{2}=\varnothing$. Since $U_{1}, U_{2}$ are open, then $f^{-1}\left(U_{1}\right), g^{-1}\left(U_{2}\right)$ are both open, hence $f^{-1}\left(U_{1}\right) \cap g^{-1}\left(U_{2}\right)$ is a neighbourhood of $x$ such that $f^{-1}\left(U_{1}\right) \cap g^{-1}\left(U_{2}\right) \subseteq A$. From Proposition 12.1.19 we find that $A$ is open.

Proposition 12.4.9. Every metric space is Hausdorff.
Proof. Let $(M, d)$ be a metric space and $x, y \in M$ any distinct points. Let $r:=d(x, y)$. The open balls $B_{r / 2}(x)$ and $B_{r / 2}(y)$ are disjoint, thus $M$ is Hausdorff.

Proposition 12.4.10. Every totally ordered set endowed with the order topology is a Hausdorff space.

Proof. Let $X$ be a space endowed with the order topology. Let $x, y \in X$ be distinct points. Suppose that $x<y$. If there exists a point $z \in X$ such that $x<z<y$ then the neighbourhoods of $x$ and $y$, respectively, $U=\{u \in X: u<z\}$ and $V=\{u \in X: u>z\}$ are disjoint, that is $U \cap V=\varnothing$. Suppose there is no such middle element, then $G=\{u \in X: u<y\}=\{u \in X: u \leqslant x\}$ and $H=\{u \in X: u>x\}=\{u \in X: u \geqslant y\}$ are neighbourhoods of $x$ and $y$, respectively, and moreover $G \cap V=\varnothing$. This shows that $X$ is Hausdorff.

### 12.5 Countability

Definition 12.5.1 (First countable). Let $X$ be a topological space. We say that $X$ is first countable if for all points $p \in X$ there exists a countable basis of $X$ composed of neighbourhoods of $p$. Equivalently $\chi(p, X) \leqslant \boldsymbol{\aleph}_{0}$.

## Convergence on First Countable Spaces

Proposition 12.5.2 (Sufficient condition for Hausdorff). Let $X$ be a first countable space. Then, $X$ is Hausdorff if and only if every sequence has at most one limit

Proof. Let $x, y \in X$ be distinct points, and $\mathcal{B}$ be a neighbourhood basis at $x$ and $\mathcal{A}$ be a neighbourhood basis at $y$. If $X$ is not Hausdorff, then for all $n \in \mathbf{N}$, choose a point $x_{n} \in B_{n} \cap A_{n}$, where $B_{n} \in \mathcal{B}$ and $A_{n} \in \mathcal{A}$ and consider the sequence $\left\{x_{n}\right\}_{n \in \mathbf{N}}$. Then there exists sub-sequences $\left\{x_{k}\right\}$ and $\left\{x_{j}\right\}$ for which $x_{k} \rightarrow x$ and $x_{j} \rightarrow y$.
$\square$
Definition 12.5.3 (Nested neighbourhood basis). Let $X$ be a topological space and a point $p \in X$. The infinite sequence of neighbourhoods of $p$, namely $\left(U_{i}\right)_{i \in \mathbf{N}}$ is said to be a nested neighbourhood basis at $p$ if for all $i \in \mathbf{N}, U_{i+1} \subseteq U_{i}$ and for all neighbourhood $V$ of $p$, there exists $i \in \mathbf{N}$ for which $U_{i} \subseteq V$.

Lemma 12.5.4. Let $X$ be a first countable topological space. Then, for all points $p \in X$ there exists a nested neighbourhood basis at $p$.

Proof. Since $X$ is first countable, let $\mathcal{V}$ be a countable basis for the topology of $X$ at $p$. If $|\mathcal{V}|<\infty$ then define $U_{i}:=V_{1} \cap \cdots \cap V_{|\mathcal{V}|}$ for all $i \in \mathbf{N}$. If $|\mathcal{V}|$ is infinite, then define for all $i \in \mathbf{N}$ the set $U_{i}:=V_{1} \cap \cdots \cap V_{i}$. For both cases, the sequence $\left(U_{i}\right)_{i \in \mathbf{N}}$ is a nested neighbourhood basis.

Definition 12.5.5 (Eventually in). Let $X$ be a topological space, and a sequence $\left(x_{i}\right)_{i \in \mathbf{N}} \subseteq$ $X$, and a set $A \subseteq X$. We say that $\left(x_{i}\right)_{i \in \mathbf{N}}$ is eventually in $A$ if $x_{i} \in A$ for all but finitely many $i \in \mathbf{N}$.

Lemma 12.5.6 (Sequence lemma). Let $X$ be a first countable space, and a set $A \subseteq X$, and a point $p \in X$. Then
(a) $p \in \mathrm{Cl} A$ if and only if $p$ is a limit of points of $A$.
(b) $A$ is closed in $X$ if and only if $A$ contains every limit point of sequences in $A$.
(c) $p \in \operatorname{Int}(A)$ if and only if all sequences that converge to $p$ are eventually in $A$.
(d) $A$ is open in $X$ if and only if every sequence in $X$ converging to a point of $A$ is eventually in $A$.

Proof. (a) $(\Rightarrow)$ Let $p \in \mathrm{Cl} A$, then for all neighbourhoods $V \subseteq A$ of $p$, the set $V \cap(A \backslash\{p\})$ is non-empty. Since $X$ is first countable, consider $\left(U_{i}\right)_{i \in \mathbf{N}}$ a nested neighbourhood basis at $p$ and construct the sequence $x: \mathbf{N} \rightarrow \bigcup_{i \in \mathbf{N}} U_{i} \cap(A \backslash\{p\})$ defined as $i \mapsto x_{i} \in$ $U_{i} \cap(A \backslash\{p\})$. We'll show that $x_{i} \rightarrow p$. Consider $V \subseteq X$ any neighbourhood of $p$, since $\left(U_{i}\right)_{i \in \mathbf{N}}$ is a basis then there exists an indexing set $I_{V} \subseteq \mathbf{N}$ for which $V=\bigcup_{i \in I_{V}} U_{i}$. Consider the index $n:=\min \left(I_{V}\right)$, then, from the definition of the nested basis, we have for all $i \geqslant n$ the elements $x_{i} \in V \cap A$, hence $x_{i} \rightarrow p$. $(\Leftarrow)$ Suppose that $\left(x_{i}\right)_{i \in \mathbf{N}}$ is a sequence of points in $A$ such that $x_{i} \rightarrow p$. From definition, for all neighbourhood $V \subseteq X$ of $p$, there exists $n \in \mathbf{N}$ such that $\forall i \geqslant N, x_{i} \in V$, moreover, since $x_{i} \in A$ then $x_{i} \in V \cap A$ which implies that $p \notin \operatorname{Ext}(A)$, hence $p \in \mathrm{Cl} A$.
(b) $(\Rightarrow)$ Suppose that $A$ is closed, then $A=\mathrm{Cl} A$. Consider any sequence $\left(x_{i}\right)_{i \in \mathbf{N}}$ of points in $A$ and let $x_{i} \rightarrow p$. From item (a) we have that $p \in A$. ( $\Leftarrow$ ) Suppose the contrary, then given any $p \in \mathrm{Cl} A$ we have $p \in A$, which implies that $A$ is closed.
(c) $(\Rightarrow)$ Suppose $p \in \operatorname{Int}(A)$. Let $\left(x_{i}\right)_{i \in \mathbf{N}} \subseteq X$ be a sequence such that $x_{i} \rightarrow p$. Consider any neighbourhood $V \subseteq A$ of $p$, then from the definition of convergence, exists $N \in \mathbf{N}$ such that $\forall i \geqslant N, x_{i} \in V_{p}$, then there exists at most $N-1$ points of $\left(x_{i}\right)_{i \in \mathbf{N}}$ outside $A$, hence the sequence is eventually in $A$. $(\Leftarrow)$ Suppose $\left(x_{i}\right)_{i \in \mathbf{N}}$ is not-eventually in $A$ and $x_{i} \rightarrow p$. For the sake of contradiction, suppose that $p \in \operatorname{Int}(A)$, then given a neighbourhood $V \subseteq A$ of $p$, there must exist $N \in \mathbf{N}$ such that $\forall i \geqslant N, x_{i} \in V \subseteq A$, which is a contradiction to the fact that $\left(x_{i}\right)_{i \in \mathbf{N}}$ is not-eventually in $A$, hence $p \notin \operatorname{Int}(A)$.
$(\mathrm{d})(\Rightarrow)$ Let $A$ be open, then $A=\operatorname{Int}(A)$. Given any $p \in A$ we have from item (c) that $p$ is the limit of all of its converging sequences are eventually in $A$. ( $\Leftarrow)$ Suppose the contrary, and let $p \in A$ be any point. From item (c) we see that $p \in \operatorname{Int}(A)$, which implies that $A \subseteq \operatorname{Int}(A)$ and hence $A$ is open.

## Second Countable Spaces and Covers

Definition 12.5.7 (Second countable). A topological space $X$ is second countable if it admits a countable basis for its topology. Equivalently $w(X) \leqslant \boldsymbol{\aleph}_{0}$.

Definition 12.5.8 (Cover). Let $X$ be a topological space and $\mathcal{U} \subseteq 2^{X}$. We say that $\mathcal{U}$ is a cover of $X$ if for all points $x \in X$ there exists $U \in \mathcal{U}$ such that $x \in U$.

Definition 12.5.9 (Open and closed covers). Let $\mathcal{U}$ be a cover of a space $X$. If all elements $U \in \mathcal{U}$ are opens in $X$, then $\mathcal{U}$ is said to be open. On the other hand, if all $U \in \mathcal{U}$ are closed in $X$, then $\mathcal{U}$ is said to be closed.

Definition 12.5.10 (Subcover). Let $\mathcal{U}$ be a cover of a space $X$. If $\mathcal{U}^{\prime} \subseteq \mathcal{U}$ is a cover for $X$, then we call it a subcover of $\mathcal{U}$.

Although formally we should always write down if the cover is either open or closed, I'll slip here and there and every time that I use the term "cover" and do not specify its type, it should be understood that I'm talking about open covers otherwise I'll specifically write "closed cover" in order to avoid any confusion.

Proposition 12.5.11 (Basis out of covers). Let $\mathcal{U}$ be an open cover of the space $X$. For all $U \in \mathcal{U}$, define $\mathcal{B}_{U}$ as the basis for the subspace $U$. Then union $\bigcup_{U \in \mathcal{U}} \mathcal{B}_{U}$ is a basis for $X$.

Proof. Define $\mathcal{B}:=\bigcup_{U \in \mathcal{U}} \mathcal{B}_{U}$. Clearly $\bigcup_{B \in \mathcal{B}} B=\bigcup_{U \in \mathcal{U}} \bigcup_{B \in \mathcal{B}} B=X$ - since every element of $X$ can be found in $\mathcal{U}$ and $\mathcal{B}_{U} \subseteq 2^{U} \subseteq 2^{X}$. Let $x \in X$ be any point and let $U \in \mathcal{U}$ be a neighbourhood of $x$. Notice that on $U$ the basis $\mathcal{B}_{U}$ satisfies the local intersecting condition (see Item 2). Since $x \in X$ is any point, then the condition is true globally for $\mathcal{B}$, thus $\mathcal{B}$ is a basis for $X$.

Corollary 12.5.12 (Second countable out of a cover). Let $\mathcal{U}$ be a countable open cover of the space $X$. If every $U \in \mathcal{U}$ is second countable, then $X$ is second countable.

Proof. We consider $\mathcal{B}$ as in Proposition 12.5.11, the union of the basis $\mathcal{B}_{U}$ for the cover elements $U \in \mathcal{U}$ - and since $U$ is second countable, we choose $\mathcal{B}_{U}$ to be a countable cover. Since $\mathcal{U}$ is countable and so is every basis contained in the basis $\mathcal{B}$, we conclude that $\mathcal{B}$ itself is countable and by Proposition 12.5 .11 we find that $\mathcal{B}$ is a basis for $X$. $\quad$.

Definition 12.5.13 (Lindelöf space). A space $X$ is said to be Lindelöf if every open cover of $X$ has a countable subcover.

Definition 12.5.14 (Separable space). A space $X$ is said to be separable if it contains a countable dense subset.

Proposition 12.5.15 (Properties of second countable spaces). The following properties hold
(SC1) A second countable space is first countable.
(SC2) A second countable space is separable.
(SC3) A second countable space is Lindelöf.
Proof. Let $\mathcal{B}$ be a countable basis for $X$.(SC1) Let a point $p \in X$, then the set $\mathcal{B}_{p} \subseteq \mathcal{B}$ of neighbourhoods of $p$ is a countable basis for $X$ at $p$.
(SC2) Let $f: I \rightarrow \bigcup_{B \in \mathcal{B}} B$ with $n \mapsto x_{n}$, where $|I|=|B|$ is an indexing set, and $x \in X$ be any point, and any neighbourhood of $V_{x} \subseteq X$ of $x$. Since $B$ is a basis, there exists an indexing set $I_{V_{x}}$ such that $V_{x}=\bigcup_{i \in I_{V_{x}}} B_{i}$, hence $V_{x} \cap A$ is non-empty. Moreover, define a sequence $\left(x_{i}\right)_{i \in I_{V_{x}}}$ such that $x_{i} \in B_{i}$. Notice that clearly $x_{i} \rightarrow x$, from the fact that $V_{x}$ is a neighbourhood of $x$, and $x_{i} \in A$, from Lemma 12.5 .6 we have that $x \in \mathrm{Cl}(\operatorname{im}(f))$. Hence we conclude that $X=\mathrm{Cl}(\operatorname{im}(f))$ and thus $\operatorname{im}(f)$ is a countable dense subset of X.
(SC3) Let $\mathcal{U}$ be a cover for $X$ and define $\mathcal{B}^{\prime}:=\{B \in \mathcal{B}: B \subseteq \mathcal{U}(U \in \mathcal{U})\}$, and $\mathcal{U}^{\prime}:=\left\{U \in \mathcal{U}: B \subseteq U\left(B \in \mathcal{B}^{\prime}\right)\right\}$. We'll show that $\mathcal{U}^{\prime}$ is a subcover of $\mathcal{U}$. Let $x \in X$ be any point. since $\mathcal{U}$ covers $X$ then there exists $U_{x} \in \mathcal{U}$ such that $x \in U_{x}$. Moreover, since $\mathcal{B}$ is a basis for the topology of $X$, then there exists $B_{x} \in \mathcal{B}$ such that $B_{x} \subseteq U_{x}$, so that $B_{x} \in \mathcal{B}^{\prime}$ and $U_{x} \in \mathcal{U}^{\prime}$ from the construction of the collections. Hence $\mathcal{U}^{\prime}$ covers X.

Proposition 12.5.16. Given a metric space $M$, the following properties are equivalent (MS1) $M$ is second countable.
(MS2) $M$ is separable.
(MS3) $M$ is Lindelöf.
Proof. (Separable $\Rightarrow$ Lindelöf) Suppose $M$ is separable and let $A$ be a countable dense set in $M$. Let $\mathcal{U}$ be a cover of $M$ and define the collection $\mathcal{U}^{\prime}:=\{U \in \mathcal{U}: U \supseteq$ $B(a, r),(a, r) \in A \times \mathbf{Q}\}$, where $B(a, r)$ is the open ball of radius $r$ around $a$. We show that $\mathcal{U}^{\prime}$ is a subcover of $M$. Let $x \in M$ be any point. Since $\mathcal{U}$ covers $M$ then there exists $U_{x} \in \mathcal{U}$ such that $x \in U_{x}$. Since $M$ is a metric space, there exists an open ball
$B(x, \ell) \subseteq U_{x}$ for some $\ell \in \mathbf{R}$. Since $A$ is dense in $M$, there exists a point $a \in A$ such that $a \in B(x, \ell / 2)$. Now, from the fact that $\mathbf{Q}$ is dense in $\mathbf{R}$ we conclude that there exists an $r \in \mathbf{Q}$ such that $d(x, a)<r<\ell / 2$. Notice that $B(a, r) \subseteq B(x, r) \subseteq U_{x}$, hence $U_{x} \in \mathcal{U}^{\prime}$. Therefore $\mathcal{U}^{\prime}$ is a subcover of $\mathcal{U}$, and since $\left|\mathcal{U}^{\prime}\right|=|A \times \mathbf{Q}|$, we find that $\mathcal{U}^{\prime}$ is countable.
(Lindelöf $\Rightarrow$ second countable) Suppose $M$ is Lindelöf, then given a cover $\mathcal{U}:=$ $\left\{B\left(x, \frac{1}{n}\right) \subseteq M: x \in M, n \in \mathbf{N}\right\}$ of $M$ there exists a countable subcover $\mathcal{U}^{\prime}$. Let $U \subseteq M$ be any open set and let $x \in U$ be any point. Since $M$ is a metric space, there exists $r>0$ such that $B(x, r) \subseteq U$. Define now $n \in N$ such that $\frac{1}{n}<\frac{r}{2}$. Since $\mathcal{U}^{\prime}$ is a cover of $M$, there exists a $B\left(p, \frac{1}{n}\right) \in \mathcal{U}^{\prime}$ such that $x \in B\left(p, \frac{1}{n}\right)$. We now show that $B\left(p, \frac{1}{n}\right) \subseteq B(x, r)$. Let $y \in B\left(p, \frac{1}{n}\right)$ be any point, then $d(p, y)<\frac{1}{n}$, moreover $d(x, p)<\frac{1}{n}$, hence we conclude that $d(x, y) \leqslant d(x, p)+d(p, y)<\frac{2}{n}<r$. This implies in particular that $y \in B(x, r)$ and in general that $B\left(p, \frac{1}{n}\right) \subseteq B(x, r) \subseteq U$. Hence we conclude that $U$ can be written as a union of elements of $\mathcal{U}^{\prime}$, which implies that $\mathcal{U}^{\prime}$ is a countable basis for $M$.

By means of Proposition 12.5.15 we conclude the equivalence chain.
Corollary 12.5.17. Euclidean spaces are countable.
Proof. Lets consider $\mathbf{R}^{n}$, notice that $\mathbf{Q}^{n}$ is dense in $\mathbf{R}^{n}$ and countable, thus $\mathbf{R}^{n}$ is separable, which by Proposition 12.5.16 implies that $\mathbf{R}^{n}$ is second countable.

## Weights and Cardinality

Proposition 12.5.18. Let $X$ be a topological space and $w(X) \leqslant \mathbf{m}$. Then for every collection $\left\{U_{i}\right\}_{i \in I} \subseteq 2^{X}$ of open sets, there exists $I_{0} \subseteq I$ such that $\left|I_{0}\right| \leqslant \mathbf{m}$ and $\bigcup_{i \in I_{0}} U_{i}=$ $\bigcup_{i \in I} U_{i}$.

Proof. Since $I_{0} \subseteq I$ then clearly $\bigcup_{i \in I_{0}} U_{i} \subseteq \bigcup_{i \in I} U_{i}$. Let $\mathcal{B}$ be a base for $X$ such that $|\mathcal{B}| \leqslant \mathbf{m}$ and define the collection $\mathcal{B}_{0}:=\left\{U \in \mathcal{B}: U \subseteq U_{i}(i \in I)\right\}$. Define a function $f: \mathcal{B}_{0} \rightarrow I$ such that for all $U \in \mathcal{B}_{0}$ we have $U \subseteq U_{f(U)} \in\left\{U_{i}\right\}_{i \in I}$. Define the indexing set $I_{0}:=f\left(\mathcal{B}_{0}\right) \subseteq I$. Notice that $\left|I_{0}\right| \leqslant|\mathcal{B}| \leqslant \mathbf{m}$. For any point $x \in \bigcup_{i \in I} U_{i}$ there exists $i \in S$ such that $x \in U_{i}$, and hence exists $U \in \mathcal{B}$ such that $x \in U \subseteq U_{i}$, from the fact that $\mathcal{B}$ is a basis. From the definition of $\mathcal{B}_{0}$ it follows that $U \in \mathcal{B}_{0}$ and hence $f(U) \in I_{0}$. Therefore, from the construction of $f$ it follows that $x \in U \subseteq U_{f(U)} \subseteq \bigcup_{i \in I_{0}} U_{i}$. From this we conclude that $\bigcup_{i \in I} U_{i} \subseteq \bigcup_{i \in I_{0}} U_{i}$.

Proposition 12.5.19. Let $X$ be a topological space. If $w(X) \leqslant \mathbf{m}$, then for every base $\mathcal{B}$ for the topology of $X$ there exists a base $\mathcal{B}_{0}$ such that $\left|\mathcal{B}_{0}\right| \leqslant \mathbf{m}$ and $\mathcal{B}_{0} \subseteq \mathcal{B}$.

Proof. Suppose $\mathbf{m} \geqslant \boldsymbol{\aleph}_{0}$. Let $\mathcal{B}:=\left\{U_{i}\right\}_{i \in I}$ be any base for the space $X$. Define a base $\mathcal{B}_{1}:=\left\{W_{t}\right\}_{t \in T}$ for $X$ such that $|T| \leqslant \mathbf{m}$.

For all $t \in T$ define the set $I(t):=\left\{i \in I: U_{i} \subseteq W_{t}\right\}$. From the fact that $W_{t}$ is open and $\mathcal{B}$ is a base for $X$, we have that $W_{t}=\bigcup_{i \in I(t)} U_{i}$. From Proposition 12.5.18 we find that there exists $I_{0}(t) \subseteq I(t)$ for which $\left|I_{0}(t)\right| \leqslant \mathbf{m}$ and

$$
\begin{equation*}
W_{t}=\bigcup_{i \in I(t)} U_{i}=\bigcup_{i \in I_{0}(t)} U_{i} \tag{12.1}
\end{equation*}
$$

Now, define the set $\mathcal{B}_{0}:=\left\{U_{i}\right\}_{i \in I(t), t \in T} \subseteq \mathcal{B}$. Since $|T|,\left|S_{0}(t)\right| \leqslant \mathbf{m}$ and the fact that $\mathbf{m}^{2}=\mathbf{m}$, we find that $\left|\mathcal{B}_{0}\right| \leqslant \mathbf{m}$. We now show that $\mathcal{B}_{0}$ is a base for the space $X$. Let a any point $x \in X$ and a neighbourhood $U \subseteq X$ of $x$. From the hypothesis that $\mathcal{B}_{1}$ is a basis, there exists $t \in T$ such that $W_{t} \subseteq U$. On the other hand Eq. (12.1) ensures the existence of an $i \in I_{0}(t)$ such that $U_{i} \subseteq W_{t} \subseteq U$. Since $U_{i} \in \mathcal{B}_{0}$, we find that $\mathcal{B}_{0}$ is indeed a basis for $X$.

### 12.6 Filters and Nets

Definition 12.6.1 (Filter on a set). A filter on a set $X$ is a collection $\mathcal{F} \subseteq 2^{X}$ that satisfies the following properties
(F1) (Downward directed) Given $A, B \in \mathcal{F}$, then there exists $C \in \mathcal{F}$ such that $C \subseteq A \cap B$.
(F2) (Upward closed) If $A \in \mathcal{F}$ and $A \subseteq B$ then $B \in \mathcal{F}$.
(F3) $\mathcal{F}$ is non-empty.
Definition 12.6.2 (Proper filter). If $\mathcal{F}$ is a filter of $X$ such that there exists $A \subseteq X$ for which $A \notin \mathcal{F}$, then we say that $\mathcal{F}$ is a proper filter. This is equivalent of saying that $\varnothing \notin \mathcal{F}$.

Proposition 12.6.3 (Eventuality filter). Let $X$ be a topological space and $\left(x_{i}\right)_{i \in \mathbf{N}}$ be a sequence of points of $X$. The collection of sets such that $\left(x_{i}\right)$ is eventually in, explicitly

$$
\mathcal{E}_{\left(x_{i}\right)}=\left\{U \subseteq X: \forall i \geqslant N, x_{i} \in U(N \in \mathbf{N})\right\},
$$

is a filter on $X$.
Definition 12.6.4 (Filter base). A non-empty downward directed set is called a filter base.

Proposition 12.6.5. Given a filter base $\mathcal{G}$ on $X$, we define the filter generated by $\mathcal{G}$ as the collection

$$
\mathcal{G}^{\uparrow}=\{A \subseteq X: G \subseteq A(G \in \mathcal{G})\}
$$

Proof. Since $\mathcal{G}$ is non-empty, $\mathcal{G}^{\uparrow}$ is non-empty. Let $A, B \in \mathcal{G}^{\uparrow}$, then there are $G, H \in \mathcal{G}$ such that $G \subseteq A$ and $H \subseteq B$. Since $G \cap H \in \mathcal{G}$, then $G \cap H \subseteq A \cap B$ and thus $A \cap B \in \mathcal{G}^{\uparrow}$. Suppose that $A \subseteq C$ where $C \subseteq X$, then since $G \subseteq A$, it follows that $G \subseteq C$ and hence $C \in \mathcal{G}^{\uparrow}$. Therefore $\mathcal{G}^{\uparrow}$ satisfies all of the three requirements.

Proposition 12.6.6. Let $X$ be a topological space. The base at a point $p \in X$, namely $\mathcal{B}_{p}$, is a filter base.

Proof. Let $U, V \in \mathcal{B}_{p}$ be neighbourhoods of $p$. Since $p \in U \cap V$ and the finite intersection of open sets is open, then $U \cap V$ is a neighbourhood of $p$ and hence $U \cap V \in \mathcal{B}_{p}$.

Definition 12.6.7 (Convergence of filters). Let $\mathcal{F}$ be a filter on a topological space $(X, \tau)$. We say that $\mathcal{F}$ converges to $x, \mathcal{F} \rightarrow x$, if and only if $\mathcal{B}_{x} \subseteq \mathcal{F}$.

## Filters determine Hausdorff, closure and continuity

Proposition 12.6.8 (Hausdorff). A topological space is Hausdorff if and only if limits of convergent proper filters are unique.

Proof. $(\Rightarrow)$ Let $X$ be a Hausdorff space and $\mathcal{F}$ be a proper filter on $X$. Suppose that for distinct $x, y \in X$ we have $\mathcal{F} \rightarrow x$ and also $\mathcal{F} \rightarrow y$. Since $X$ is Hausdorff, let $U, V \subseteq X$ neighbourhoods of $x$ and $y$, respectively, and such that $U \cap V$ is empty. Notice that since $U$ is a neighbourhood of $x$ then $U \in \mathcal{F}$, and analogously, $V \in \mathcal{F}$, but from the downward directness, it follows that $U \cap V=\varnothing \in \mathcal{F}$, which can't be the case since $\mathcal{F}$ is supposed to be proper. Hence $\mathcal{F}$ cannot converge to distinct points of $X$. $(\Leftarrow)$ Suppose that $X$ is not Hausdorff and choose distinct points $x, y$ that are not separable by open sets. Consider the collection $\mathcal{B}=\mathcal{B}_{x} \cap \mathcal{B}_{y}$, then given two sets $A, B \in \mathcal{B}$ we have that $A \cap B \in \mathcal{B}$ and therefore $\mathcal{B}$ is a filter base. Notice that the filter $\mathcal{B}^{\uparrow}$ converges both to $x$ and $y$.

Proposition 12.6.9 (Closed). Let $X$ be a topological space and $A \subseteq X$. A point $p \in \mathrm{Cl} A$ if and only if there exists a proper filter $\mathcal{F}$ with $A \in \mathcal{F}$ such that $\mathcal{F} \rightarrow p$.

Proof. $(\Rightarrow)$ Let $p \in \mathrm{Cl} A$, then for all neighbourhoods $U \in \mathcal{B}_{p}$ the set $U \cap(A \backslash\{p\})$ is non-empty, in particular we have that $\mathcal{B}:=\left\{U \cap A: U \in \mathcal{B}_{p}\right\}$ does not contain the empty set. Hence the proper filter $\mathcal{B}^{\uparrow}$ converges to $p$. $(\Leftarrow)$ Let $\mathcal{F}$ be a proper filter with $\mathcal{F} \rightarrow p$ and $A \in \mathcal{F}$, then in particular we have that the downward directness implies $\mathcal{B} \subseteq \mathcal{F}$ and thus $\varnothing \notin \mathcal{B}$, hence $x \in \mathrm{Cl} A$.

Definition 12.6.10 (Pushforward of filters). Let $f: X \rightarrow Y$ be a map of sets. The collection of images $\{f(A): A \in \mathcal{F}\}$ form a filter base whose generated filter is defined to the be the pushforward of $\mathcal{F}$ with respect to $f$, namely $f_{*}(\mathcal{F})$. Hence

$$
f_{*}(\mathcal{F})=\{B \subseteq Y: f(A) \subseteq B(A \in \mathcal{F})\}
$$

Proposition 12.6.11 (Continuity). A map $f: X \rightarrow Y$ is continuous if and only if for every given filter $\mathcal{F}$ on $X$ such that $\mathcal{F} \rightarrow x$, then $f_{*}(\mathcal{F}) \rightarrow f(x)$, where $x \in X$.

Proof. $(\Rightarrow)$ Suppose $f$ is a continuous map and $\mathcal{F} \rightarrow x$. From continuity of $f$, for any given neighbourhood of $f(x), B \in \mathcal{B}_{f(x)}$, there exists a corresponding neighbourhood of $x, V \in \mathcal{B}_{x}$, such that $f^{-1}(B) \subseteq V$. Since $\mathcal{F} \rightarrow x$ if and only if $\mathcal{B}_{x} \subseteq \mathcal{F}$ (from Definition 12.6.7), then in particular $f^{-1}(B) \in \mathcal{F}$, since $f^{-1}(B) \in \mathcal{B}_{x}$, is a neighbourhood of $x$. Hence, if $B \in \mathcal{B}_{f(x)}$ is any element and $A=f^{-1}(B)$, then $f(A) \subseteq B$, which implies that $\mathcal{B}_{f(x)} \subseteq f_{*}(\mathcal{F})$ from the definition of the pushforward, therefore $f_{*}(\mathcal{F}) \rightarrow f(x)$.
$(\Leftarrow)$ Suppose now that for any filter $\mathcal{F}$ on $X$ such that $\mathcal{F} \rightarrow x$, implies $f_{*}(\mathcal{F}) \rightarrow f(x)$. Given any open set $U \subseteq Y$, if $U \cap \operatorname{im}(f)=\varnothing$ then $f^{-1}(U)=\varnothing$ and hence is open, otherwise there exists $x \in X$ for which $f(x) \in U$. Given such a point $x \in X$, take $\mathcal{F}=\mathcal{B}_{x}^{\uparrow}$ so that from hypothesis $f_{*}(\mathcal{F}) \rightarrow f(x)$ and hence $\mathcal{B}_{f(x)} \subseteq f_{*}(\mathcal{F})$. Since $U$ is open, then $U \in \mathcal{B}_{f(x)}$, which in turn implies that $U \in f_{*}(\mathcal{F})$ and from definition there must exist a set $V \in \mathcal{F}$, which happens to be a neighbourhood of $x$, such that $f(V) \subseteq B$, thus $f$ is continuous.

## Write on nets later

### 12.7 Topological Manifolds

## Locally Euclidean

Definition 12.7.1 (Locally euclidean). Let $M$ be a topological space. We say that $M$ is locally euclidean of dimension $n$ if for all points $x \in X$ there exists a neighbourhood $U_{x} \subseteq M$ of $x$ such that exists an open set $V \subseteq \mathbf{R}^{n}$ for which $U_{x} \simeq V$, that is, $U_{x}$ is isomorphic to $V$.

Proposition 12.7.2. Let $B^{n} \subseteq \mathbf{R}^{n}$ be any open ball, then the morphism

$$
\varphi: B^{n} \rightarrow \mathbf{R}^{n}, x \longmapsto \frac{x}{1-|x|}
$$

is an isomorphism $B^{n} \simeq \mathbf{R}^{n}$.
Lemma 12.7.3. A topological space $M$ is locally euclidean of dimension $n$ if and only if either of the following is true:
(a) Every point of $M$ has a neighbourhood in $M$ that is isomorphic to an open ball in $\mathbf{R}^{n}$.
(b) Every point of $M$ has a neighbourhood in $M$ that is isomorphic to $\mathbf{R}^{n}$.

Proof. From Proposition 12.7.2, we find that the proof of one of the two statements proves the other. Lets prove the first one of them. $(\Leftarrow)$ Since $B^{n} \subseteq \mathbf{R}^{n}$ is an open of $\mathbf{R}^{n}$ we conclude from the definition Definition 12.7.1 that $M$ is indeed locally euclidean. $(\Rightarrow)$ Suppose $M$ is a locally euclidean $n$-dimensional topological space, then let any point $p \in M$ and define $U_{p} \subseteq M$ be a neighbourhood of $p$ such that exists $V \subseteq \mathbf{R}^{n}$ for which we can define an isomorphism $\phi: U_{p} \rightarrow V$. From the fact that the collection of open balls in $\mathbf{R}^{n}$ form a basis for $\mathbf{R}^{n}$, we conclude that there exists an open ball $B^{n} \subseteq V$ for which $p \in B^{n}$. Thus the open set $\psi^{-1}(B) \subseteq M$ is a neighbourhood of $p$. Since $\phi$ is an isomorphism, then the morphism $\left.\psi\right|_{B} ^{-1}: B \rightarrow \phi^{-1}(B)$ is an isomorphism, thus $B \simeq \phi^{-1}(B)$ as wanted.

Proposition 12.7.4. Every locally euclidean space is first countable.

## prove

Definition 12.7.5 (Miscellaneous). Let $M$ be a locally euclidean $n$-dimensional topological space. We define:
(C1) (Coordinate domain) An open set $U \subseteq M$ is called a coordinate domain of $M$ if it is isomorphic to an open set of $\mathbf{R}^{n}$.
(C2) (Coordinate map) An isomorphism $\phi$ from a coordinate domain to an open set of $\mathbf{R}^{n}$ is called a coordinate map.
(C3) (Coordinate chart) The pair $(U, \phi)$ of a coordinate domain and one of its coordinate maps is called a coordinate chart for $M$. Given $x \in M$, we say that $(U, \phi)$ is a chart at $x$ if $x \in U$.
(C4) (Coordinate ball) A coordinate domain isomorphic to an open ball of $\mathbf{R}^{n}$ is called a coordinate ball.
(C5) (Coordinate neighbourhood or euclidean neighbourhood) Given a point $p \in M$, if $U \subseteq M$ is a coordinate domain of $M$ such that $p \in U$, then we say that $U$ is a coordinate neighbourhood of $p$.

Proposition 12.7.6. Every coordinate ball is second countable.
Proof. Let $M$ be a locally euclidean $n$-dimensional topological space and let $U \subseteq M$ be a coordinate ball of $M$ and $\phi: U \xrightarrow{\simeq} B^{n}$ be an isomorphism. From Corollary 12.5.17 we find that there exists a countable basis $\mathcal{B} \subseteq 2^{\mathrm{R}^{n}}$ for $B^{n}$. Since $\phi$ is an isomorphism, then the collection of preimages $\left\{\phi^{-1}(B) \subseteq U: B \in \mathcal{B}\right\}$ is also a basis for $U$ and is clearly countable - hence $U$ is second countable.

Proposition 12.7.7 (Locally euclidean from surjective morphism). Let $X$ be locally euclidean of dimension $n$, and $f: X \rightarrow Y$ be a surjective local isomorphism. Then $Y$ is locally euclidean of dimension $n$.

Proof. Let $\mathcal{B}$ be a basis for $X$. Since $f$ is continuous, surjective and a open (see Proposition 12.2.14), we can use Proposition 12.2.7 to conclude that $f(\mathcal{B})$ is a basis for $Y$. Since $f$ is a local isomorphism, given any point $y \in Y$, choose a neighbourhood $V_{y} \in f(\mathcal{B})$, so that there exists a $B \in \mathcal{B}$ such that $f(B)=V_{y}$. Consider the isomorphism $\left.f\right|_{B}: B \rightarrow V_{y}$. Since $X$ is locally euclidean of dimension $n$, there exists an open set $W \subseteq \mathbf{R}^{n}$ such that $B \simeq W$. Moreover, since $B \simeq V_{y}$, then $V_{y} \simeq W$. We conclude that $Y$ is locally euclidean of dimension $n$.

## Topological Manifold

Definition 12.7.8 (Topological manifold). An $n$-dimensional topological manifold is a second countable Hausdorff space that is locally euclidean $n$-dimensional.

Proposition 12.7.9. Every topological manifold admits a basis of coordinate balls.
Proof. Let $M$ be a $n$-dimensional topological manifold. From Lemma 12.7.3 we can take, for every point $p \in M$, a neighbourhood $U \subseteq M$ such that $U \simeq B^{n}$ (an open ball of $\mathbf{R}^{n}$ ). Define $\mathcal{U}$ to be the collection of all coordinate balls on $M$, from our last argument it follows that $\mathcal{U}$ covers $M$. Let $U, U^{\prime} \in \mathcal{U}$ be intersecting coordinate balls and $\phi$ be a coordinate isomorphism of either $U$ or $U^{\prime}$. Let $p \in U \cap U^{\prime}$ be any point and define a ball $B_{p}^{n} \subseteq \phi\left(U \cap U^{\prime}\right) \subseteq \mathbf{R}^{n}$ that is a neighbourhood of $\phi(p) \in \mathbf{R}^{n}$. Let $V \subseteq M$ be defined as $V:=\phi^{-1}\left(B^{n}\right)$ so that $V \subseteq U \cap U^{\prime}$ and also $p \in V$. Notice that the induced
map $\psi: V \rightarrow B^{n}$, defined by $\psi(x):=\phi(x)$, is an isomorphism - that is, $V \in \mathcal{U}$ and hence $\mathcal{U}$ is a basis for $M$ (see Proposition 12.1.5).

Proposition 12.7.10. Every open subset of an $n$-manifold is itself an $n$-manifold.
Proof. Consider $U \subseteq M$ an open set and let $p \in U$. Then consider $V \subseteq M$ to be a coordinate neighbourhood of $x$ such that $V \simeq B \subseteq \mathbf{R}^{n}$, then the set $U \cap V$ is open and isomorphic to a subset of $B$ and hence $V$ is $n$-dimensional locally euclidean. Moreover, since a subset of a Hausdorff space is Hausdorff and a subset of a second countable space is second countable, then $U$ is indeed a $n$-manifold.

Definition 12.7.11. The empty topological space is an $n$-manifold for all $n>0$.
Theorem 12.7.12 (Dimension invariance). If $m \neq n$, then a non-empty topological space cannot be both $n$-manifold and $m$-manifold.

Proposition 12.7.13. A separable metric space that is locally euclidean of dimension $n$ is an $n$-manifold.

Proof. From Proposition 12.5.16, since $M$ is separable then it is also second countable. Moreover, from Proposition 12.4.9 we find that $M$ is Hausdorff. Hence $M$ is an $n$ manifold.

Proposition 12.7.14. Every topological manifold is separable and metrizable.
Proof. Notice that since a manifold is second countable, then by Proposition 12.5.15 we find that it is separable.

## After proving Urysohn metrization theorem, prove the metrizability property

## Manifolds with Boundary

Definition 12.7.15 (Upper half-space). We define the closed $n$-dimensional upper halfspace $\mathbf{H}^{n} \subseteq \mathbf{R}^{n}$ as

$$
\mathbf{H}^{n}:=\left\{x \in \mathbf{R}^{n}: \pi_{n}(x) \geqslant 0\right\}
$$

where $\pi_{n}$ is the projection of the $n$-th coordinate. We define the boundary of $\mathbf{H}^{n}$ as $\partial \mathbf{H}^{n}:=\left\{x: \pi_{n}(x)=0\right\}$, and the interior as $\operatorname{Int}\left(\mathbf{H}^{n}\right):=\left\{x: \pi_{n}(x)>0\right\}$.

Definition 12.7.16 (Manifold with boundary). We define an $n$-dimensional topological manifold with boundary to be a second countable Hausdorff space such that each point has a neighbourhood isomorphic to an open set of $\mathbf{R}^{n}$ or $\mathbf{H}^{n}$.

Definition 12.7.17 (Miscellaneous). Let $M$ be an $n$-manifold with boundary. We define
(MB1) A coordinate chart for $M$ is a pair $(U, \phi)$, where $U \subseteq M$ is an open set and $\phi: U \xrightarrow{\simeq} V$ is an isomorphism, where $V \subseteq \mathbf{R}^{n}$ or $V \subseteq \mathbf{H}^{n}$. We say that the chart is an interior chart if $V$ is an open subset of $\mathbf{R}^{n}$. A chart is said to be a boundary chart if $V$ is an open subset of $\mathbf{H}^{n}$ with $\operatorname{im}(\phi) \cap \partial \mathbf{H}^{n} \neq \varnothing$.
(MB2) A point $p \in M$ is called an interior point of $M$ if it is contained in the domain of an interior chart. The collection of such points is called the interior of $M$, and is denoted $\operatorname{Int}(M)$.
(MB3) A point $p \in M$ is called a boundary point of $M$ if it is in the domain of a boundary chart that maps $p$ to a point of $\mathbf{H}^{n}$. The boundary of $M$, is defined as the collection of such points and is denoted by $\partial M$.

Proposition 12.7.18. If $M$ is an $n$-dimensional manifold with boundary, then $\operatorname{Int}(M)$ is an open subset of $M$, which is itself an $n$-dimensional manifold (without boundary).

Proof. Let $p \in \operatorname{Int}(M)$ be any point, then by definition we have that $p \in U$ where $(U, \phi)$ is an interior chart for $M$ and therefore $U \simeq V$ where $V$ is some open subset of $\mathbf{R}^{n}$, hence locally euclidean, which makes $\operatorname{Int}(M)$ an $n$-dimensional manifold (since a subset of a Hausdorff space is Hausdorff and a subset of a second countable space is again second countable). To prove that $\operatorname{Int}(M)$ is open, we can use the fact that $M$ is first countable and lemma Lemma 12.5.6. First, suppose the converse, so that $\left(x_{i}\right)_{i \in \mathbf{N}}$ is a sequence of points in $M$ that is not-eventually in $\operatorname{Int}(M)$ and, for the sake of contradiction, $x_{i} \rightarrow p$. Consider the open neighbourhood $U$ of $p$ (which happens to be the chart domain), if $x_{i} \rightarrow p$ then there exists $N \in \mathbf{N}$ such that $x_{i} \in U$ for all $i \geqslant N$, but notice that every point of $U$ is an interior point of $M$, therefore $x_{i} \in \operatorname{Int}(M)$, which is a contradiction to the hypothesis that the sequence is not-eventually in $\operatorname{Int}(M)$. Hence such sequences cannot converge to $p$, and $\operatorname{Int}(M)$ is open.

Theorem 12.7.19 (Boundary invariance). Let $M$ be a manifold with boundary, then

$$
\partial M \cap \operatorname{Int}(M)=\varnothing
$$

## Proof.

Proof to come far ahead

Corollary 12.7.20. If $M$ is a non-empty $n$-manifold with boundary, then the collection $\partial M$ is closed in $M$, and $M$ is an $n$-manifold if and only if $\partial M=\varnothing$.

Proof. Since $\partial M=M \backslash \operatorname{Int}(M)$ from theorem Theorem 12.7.19 and $\operatorname{Int}(M)$ is open from Proposition 12.7.18, then it follows that $\partial M$ is closed. $(\Rightarrow)$ Moreover, suppose that $M$ is a manifold, then given any point $p \in M$ there exists an interior chart $(U, \phi)$ such that $p \in U \subseteq \operatorname{Int}(M)$, hence $\operatorname{Int}(M)=M$, and from Theorem 12.7.19 we conclude that $\partial M=\varnothing$. $(\Leftarrow)$ Suppose that $\partial M=\varnothing$, then from Theorem 12.7.19 we conclude that $M=\operatorname{Int}(M)$, which is a manifold by Proposition 12.7.18.

## Chapter 13

## Top - Universal Properties

### 13.1 Prelude

Before diving into the construction of new spaces from old using the classical categorical notions of limits over diagrams, lets first stablish the definition of initial and final topologies, that will classify the kind of topology each of these new constructions receive.

Definition 13.1.1 (Initial \& final topology). Let $S$ and $J$ be sets, and consider a collection of topological spaces $\left(X_{j}\right)_{j \in J}$. We define the following notions:
(a) Given a collection of set-functions $\left(f_{j}: S \rightarrow X_{j}\right)_{j \in J}$, we define the initial topology $\tau_{\text {initial }}$ on $S$ - induced by $\left(f_{j}\right)_{j \in J}$ - to be the minimum collection of open sets such that $f_{j}:\left(S, \tau_{\text {initial }}\right) \rightarrow X_{j}$ are continuous maps.
(b) Given a collection of set-functions $\left(g_{j}: X_{j} \rightarrow S\right)_{j \in J}$, we define the final topology $\tau_{\text {final }}$ on $S$ - induced by $\left(g_{j}\right)_{j \in J}$ - to be the maximum collection of open sets such that $g_{j}: X_{j} \rightarrow\left(S, \tau_{\text {final }}\right)$ are continuous maps.

### 13.2 Subspace

## Construction

Definition 13.2.1 (Subspace topology). Given a topological space $(X, \tau)$ and a subset $S \subseteq X$, we define the space $\left(S, \tau_{S}\right)$ as a subspace of $X$ if

$$
\tau_{S}=\{U \subseteq S: U=S \cap V, V \in \tau\}
$$

If so, we call $\tau_{S}$ the relative topology or subspace topology.
Proposition 13.2.2. Let $X$ be a topological space and $S$ be a subspace of $X$. A set $A \subseteq S$ is closed in $S$ if and only if $A=S \cap F$ for some closed set $F \subseteq X$ with respect to $X$. The closure of $A$ with respect to $S$, denoted $\widetilde{A}$, is such that $\widetilde{A}=S \cap \mathrm{Cl} A$.

Proof. $(\Rightarrow)$ Suppose $A$ is a closed set in $S$, then $S \backslash S=S \cap U$ is open and hence $U$ is some open set of $X$. Then

$$
A=S \backslash(S \backslash A)=S \backslash(S \cap U)=S \cap(X \backslash U)
$$

where $X \backslash U$ is closed, hence $A$ equals to the intersection of $S$ with a closed set of $X$. $(\Leftarrow)$ Suppose that $A=S \cap F$ for a closed set $F$ in $X$. Then we have

$$
S \backslash A=S \backslash(S \cap F)=S \cap(X \backslash F)
$$

Since $X \backslash F$ is open, then $S \backslash A$ is open. We conclude that $A$ is closed in $S$.
For the last proposition, notice that $\widetilde{A}=\bigcap\{F \subseteq S: F \supseteq A, F$ is closed in $S\}$, that is, the intersection of sets $S \cap C$ such that $C$ is closed in $X$ and $C \supseteq A$, so that $\widetilde{A}=S \cap \mathrm{Cl} A$.

Proposition 13.2.3. Let $X$ be a topological space and $S$ be a subspace. Then
(a) If $U \subseteq S$ is open (resp. closed) and $S$ is an open (resp. closed) subset of $X$, then $U$ is open (resp. closed) in $X$.
(b) If $U \subseteq S$ and $U$ is open (resp. closed) in $X$, then it is open (resp. closed) in $S$.

Proof. (a) Suppose that $U$ is open in $S$ and $S$ is open in $X$. From the definition of the relative topology, there exists an open $V \subseteq X$ in $X$ such that $U=S \cap V$. Since $S$ and $V$ are both open in $X$, then $U$ is open in $X$. From Proposition 13.2.2 we find that if $U$ is closed in $S$ then $B=S \cap F$ for some closed set $F \subseteq X$. Moreover, if $S$ is closed in $X$, then clearly $B$ is closed in $X$.
(b) If $B \subseteq S$ then clearly $B \cap S=B$, hence if $B$ is open in $X$, so it is in $S$. Moreover, if $B$ is closed in $X$, from Proposition 13.2.2 $B$ is closed in $S$.

Definition 13.2.4 (Second definition of the subspace topology). Let $X$ be a topological space and $S \subseteq X$ be any set. The subspace topology on $S$ is the initial topology on the set such that the inclusion $\tau_{s}: S \rightarrow X$ is continuous.

Corollary 13.2.5. Definitions 13.2 .1 and 13.2.4 are equivalent.
Proof. Suppose $\tau_{\iota S}$ is the initial topology such that $\tau_{S}$ is continuous and $\tau_{S}$ be the subspace topology. Our goal is to show that they are, in fact, equal. Given any $U=S \cap V \in \tau_{S}$ we have $\iota_{S}^{-1}(V)=S \cap V=U$, since $V$ is open, we find that $U$ is open in $\tau_{l S}$. This implies in $\tau_{S} \subseteq \tau_{l S}$. Moreover, let $O \in \tau_{l S}$, then there must exist some $A \subseteq X$ open set for which $\iota_{S}^{-1}(A)=A \cap S=O$ (this comes directly from the fact that $\tau_{l S}$ was solely constructed for the purpose of making $\iota_{S}$ continuous), since $A$ is open, then $O=A \cap S \in \tau_{S}$. This implies that $\tau_{\iota S} \subseteq \tau_{S}$. We conclude finally that both definitions are indeed equivalent.

Theorem 13.2.6 (Universal property of the subspace topology). Let $X$ be a topological space and $S$ be a subspace. Given any topological space $Y$, a map $f: Y \rightarrow S$ is continuous
if and only if $\iota_{S} f: Y \rightarrow X$ is continuous, where $\iota_{S}: S \hookrightarrow X$ is the inclusion map. Hence the following diagram commutes


Moreover, on the converse, if $S \subseteq X$ is a topological space such that the above property holds, then it is equipped with the subspace topology.

Proof. First we show that if $\left(S, \tau_{S}\right)$ is a subspace of $(X, \tau)$ then it satisfies the universal property. Suppose then that $S$ a subspace of $X .(\Rightarrow)$ Let $f$ be continuous. If $V \subseteq X$ is any open subset, we have that

$$
(\iota S f)^{-1}(V)=f^{-1}\left(\iota^{-1}(V)\right)=f^{-1}(S \cap V)
$$

since $V$ is said to be open, then $S \cap V$ is open in $S$, which implies that $f^{-1}(S \cap V)=$ $\left(\iota_{S} f\right)^{-1}(V)$ is open. This shows that the map $\iota_{S} f$ is open. $(\Leftarrow)$ Let $\iota_{S} f$ be continuous. Consider $U=S \cap V=\iota_{S}^{-1}(V)$ to be any open set of the subspace $S$ (that is $V$ is an open of the space $X$ ). Then we have

$$
f^{-1}(U)=f^{-1}\left(\iota_{S}^{-1}(V)\right)=(\iota S f)^{-1}(V)
$$

since $\iota_{s} f$ is continuous, then $f^{-1}(U)$ is open, therefore $f$ is continuous.
We now show that if an object satisfies such the property, then it is the subspace. Let $\left(S, \tau^{\prime}\right)$ be a space satisfying the universal property. In particular we can take the subspace $\left(S, \tau_{S}\right)$ of $(X, \tau)$ and the identity map $\operatorname{id}_{S}:\left(S, \tau_{S}\right) \rightarrow\left(S, \tau^{\prime}\right)$. Since $\left(S, \tau^{\prime}\right)$ satisfies the universal property, we have that id ${ }_{S}$ is continuous if and only if $\iota_{S} \mathrm{id}_{S}=\iota_{S}$ is continuous. That is


We know from Definition 13.2.4 that $\iota_{S} \mathrm{id}_{S}=\iota_{S}$ is continuous for the subspace topology $\tau_{S}$, hence the universal property implies that $\mathrm{id}_{S}$ is continuous. In particular, this says that $\tau^{\prime} \subseteq \tau_{S}$. In order to show the other side of the equality, consider now the space $\left(S, \tau^{\prime}\right)$ and the identity map $\mathrm{id}_{S}^{\prime}:\left(S, \tau^{\prime}\right) \rightarrow\left(S, \tau^{\prime}\right)$ so that from the universal property of ( $S, \tau^{\prime}$ ) the map $\mathrm{id}_{S}^{\prime}$ is continuous if and only if $\iota_{S} \mathrm{id}_{S}^{\prime}=\iota_{S}$ is continuous. That is, the following diagram commutes


Notice that since $\mathrm{id}_{S}^{\prime}$ is continuous (see Proposition 12.2.3) then from the universal property the map $\iota_{S}$ is continuous on $\tau^{\prime}$. Since $\tau_{S}$ is the initial topology such that $\tau_{S}$ is continuous (see Definition 13.2.4), then clearly $\tau_{S} \subseteq \tau^{\prime}$. This finishes the proof that $\tau^{\prime}=\tau_{S}$ and hence the space $\left(S, \tau^{\prime}\right)=\left(S, \tau_{S}\right)$ is the subspace of $(X, \tau)$.
Corollary 13.2.7. Let $X, Y$ be topological spaces and $f: X \rightarrow Y$ be a continuous map. Then the following hold
(a) (Domain restriction) Let $S$ be a subspace of $X$. Then $\left.f\right|_{S}$ is continuous.
(b) (Codomain restriction) Let $T$ be a subspace of $Y$ such that $f(X) \subseteq T$. Then $f: X \rightarrow T$ is continuous.
(c) (Codomain expansion) Let $Y$ be a subspace of $Z$. Then the map $f: X \rightarrow Z$ is continuous.

Proof. (a) Notice that $\left.f\right|_{S}=f \iota_{S}$. Applying the universal property, we find that since $f$ is continuous, so is $\left.f\right|_{S}$. (b) From the universal property we have that $\iota_{T} f=f$ is continuous, so is $f: X \rightarrow T$. (c) Notice that from the universal property we have $\iota_{Y} f: X \rightarrow Z$ continuous, since $f$ is continuous. The following are the universal property diagrams for items (b) and (c):


Proposition 13.2.8 (Subspace properties). Let $(X, \tau)$ be a topological space and $\left(S, \tau_{S}\right)$ be a subspace of $X$. The following are properties concerning the subspace topology
(SP1) Let $T$ be a subspace of $S$. Then $T$ is a subspace of $X$.
(SP2) Let $\mathcal{B}$ be a basis for $X$. Then the collection $\mathcal{B}_{S}=\{B \cap S: B \in \mathcal{B}\}$ is a basis for $S$.
(SP3) Let $\left(p_{i}\right)_{i \in \mathbf{N}} \subseteq S$ be a sequence and $p \in S$. Then $p_{i} \rightarrow p$ in $S$ if and only if $p_{i} \rightarrow p$ in $X$.
(SP4) A subspace of a Hausdorff space is Hausdorff.
(SP5) A subspace of a first countable space is first countable.
(SP6) A subspace of a second countable space is second countable.
Proof. (SP1) Let $Z$ be some topological space, we can choose maps $f: Z \rightarrow T$ and $g=\iota_{T} f: Z \rightarrow S$ and apply the universal property on both $T$ and $S$ in order to get


Moreover, if $f$ is continuous, from the universal property of $T$ we find that $g$ is continuous, but using the universal property of $S$ that tells us that $\iota_{S} g=\iota_{T}^{\prime} f$ is continuous. The converse is true by using the same argumentation. Hence $T$ satisfies the universal property and therefore is a subspace of $X$.
(SP2) First notice that since $B \in \mathcal{B}$ is open in $X$, then $\mathcal{B}_{S} \subseteq \tau_{S}$. Let $s \in S$ be any element, then in particular we have $s \in B_{s} \in \mathcal{B}$ for some element of the basis. Then $s \in S \cap B_{s} \in \mathcal{B}_{S}$. Let $A=S \cap B_{1}, B=S \cap B_{2} \in \mathcal{B}_{S}$ and $x \in A \cap B$. Then in particular $x \in B_{1} \cap B_{2}$. Since $\mathcal{B}$ is a basis for $X$, then exists $B_{3} \in \mathcal{B}$ such that $x \in B_{3} \subseteq B_{1} \cap B_{2}$. Hence the corresponding set $C=S \cap B_{3} \in \mathcal{B}_{S}$ is such that $x \in C \subseteq A \cap B$. This proves that $\mathcal{B}_{S}$ is a basis.
(SP3) $(\Rightarrow)$ Suppose $p_{i} \rightarrow p$ in $S$, that is, for all $U_{p}=S \cap V_{p} \in \tau_{S}$, where $V_{p} \in \tau$, we have some $N \in \mathbf{N}$ such that for all $n \geqslant N$ we have $p_{n} \in U_{p}$. In particular, this implies that for each $n \geqslant N$ we get $p_{n} \in V_{p}$, hence $p_{i} \rightarrow p$ in $X$. ( $\Leftarrow$ ) Suppose that $p_{i} \rightarrow p$ in $X$. Let $V_{p} \subseteq X$ be a neighbourhood of $p$, then since $p \in S$ we have that $V_{p} \cap S \neq \varnothing$ is an element of $\tau_{S}$. Let $U_{p} \in \tau_{S}$ be any neighbourhood of $p$, then there exists $V \in \tau$ such that $U_{p}=S \cap V$ (from the definition of the subspace topology) and also $p \in V$, so that $V$ is a neighbourhood of $p$. Therefore there exists $M \in \mathbf{N}$ such that, for all $n \geqslant M$ we have $p_{n} \in S \cap V=U_{p} \subseteq V$. This implies that $p_{i} \rightarrow p$ in $S$.
(SP4) Let $X$ be Hausdorff. Let $x, y \in S$ be distinct points. In particular, there exits $A, B \in X$ neighbourhoods of $x$ and $y$, respectively, such that $A \cap B=\varnothing$. Hence the sets $U=S \cap A, V=S \cap B \in \tau_{S}$ are neighbourhoods of $x$ and $y$ respectively and since $U \subseteq A$ and $V \subseteq B$ we find that $U \cap V=\varnothing$. Hence $S$ is Hausdorff.
(SP5) Let $X$ be a first countable space. Let $p \in S$ be any point. Define a countable base $\mathcal{B}_{p}$ of neighbourhoods of $p$ for $X$. Define the, clearly countable, collection $\mathcal{B}_{p}^{\prime}=$ $\left\{S \cap B_{p}: B_{p} \in \mathcal{B}_{p}\right\} \subseteq 2^{S}$. From Item 2 we have that $\mathcal{B}_{p}^{\prime}$ is a basis for $S$. Moreover, every element of $\mathcal{B}_{p}^{\prime}$ is clearly a neighbourhood of $p$. This proves the property.
(SP6) Let $X$ be second countable and $\mathcal{B}$ be a countable basis for the topology of $X$. Define the, clearly countable, collection $\mathcal{B}^{\prime}:=\{S \cap B: B \in \mathcal{B}\}$. Since $\mathcal{B}$ is a basis for $X$, we use Item 2 to conclude that $\mathcal{B}^{\prime}$ is a basis for $S$.

Proposition 13.2.9. Let $X$ be a topological space and $A, B \subseteq X$ be subspaces such that

- Either $A$ and $B$ are both open subsets of $X$ or are both closed.
- They cover $X$, that is, $X=A \cup B$.

Then the following commutative diagram is a pushout in Top


In other words, if $f: X \rightarrow Y$ is any set-function between topological spaces, then $f$ is continuous if the restrictions $\left.f\right|_{A}$ and $\left.f\right|_{B}$ are continuous.

Proof. Since diagram Eq. (13.1) is a pushout in Set, one just has to prove that $X$ has the final topology induced by the natural inclusions $\iota_{A}: A \hookrightarrow X$ and $\iota_{B}: B \hookrightarrow X —$ that is,
$U \subseteq X$ is open if and only if $\iota_{A}^{-1}(U)=U \cap A$ is open in $A$ and $\iota_{B}^{-1}(B)=U \cap B$ is open in B.

Indeed, if $U$ is an open set of $X$, then by definition of the subspace topology we conclude that $\iota_{A}^{-1}(U)$ is open in $A$ and $\iota_{B}^{-1}(U)$ is open in $B$.

On the other hand, assume that $U \cap A$ is open in $A$, while $U \cap B$ is open in $B$. We separate the proof in two cases:

- Suppose that $A$ and $B$ are open subspaces of $X$. Since $A \cap U$ is open, there must exist an open set $V_{A} \subseteq X$ such that $\iota_{A}^{-1}\left(V_{A}\right)=V_{A} \cap A=U \cap A$ - furthermore, since $A$ and $V_{A}$ are open in $X$, it follows that $U \cap A$ is open in $X$. The same argument can be made for the subspace $B$, concluding that $U \cap B$ is also open in $X$. Finally, one notices that

$$
U=U \cap X=U \cap(A \cup B)=(U \cap A) \cup(U \cap B)
$$

which implies that $U$ is open in $X$ - since it is a union of open sets.

- Suppose that $A$ and $B$ are closed subspaces of $X$. Since $A$ and $B$ are subspaces, just as in the open subspace case, there exists open sets $V_{A}, V_{B} \subseteq X$ such that $U \cap A=V_{A} \cap A$ and $U \cap B=V_{B} \cap B$. Notice that since $A$ and $B$ are closed in $X$, then both $A \backslash U$ and $B \backslash U$ are closed in $X$ - therefore by definition both $X \backslash(A \backslash U)$ and $X \backslash(B \backslash U)$ are open in $X$. Furthermore, one can write

$$
\begin{aligned}
U & =X \backslash(X \backslash U) \\
& =X \backslash((A \cup B) \backslash U) \\
& =X \backslash((A \backslash U) \cup(B \backslash U)) \\
& =(X \backslash(A \backslash U)) \cap(X \backslash(B \backslash U))
\end{aligned}
$$

that is, $U$ is the intersection of open sets, thus $U$ is an open set of $X$.

## Topological Embeddings

Definition 13.2.10 (Embedding). Let $f: Y \leadsto X$ be a continuous injective map of topological spaces. We call $f$ an embedding when $f^{\prime}: Y \stackrel{\simeq}{\rightrightarrows} f(Y)$, for $f^{\prime}(y):=f(y)$ for all $y \in Y$, is an isomorphism.

Example 13.2.11. Let $X$ be a topological space and $S$ be a subspace of $X$. We show that $\iota_{S}: S \hookrightarrow X$ is an embedding. Notice that $\iota_{S}(S)=S$ and hence $\iota_{S}^{\prime}: S \xrightarrow{\simeq} \iota(S)=S$ is equal to the identity map id ${ }_{S}$, which clearly establishes an isomorphism.
Proposition 13.2.12. Let $f$ be a continuous injective map between topological spaces. If $f$ is either open or closed, then $f$ is an embedding.
Proof. Let $f: X \rightarrow Y$. Since $f$ is injective, the codomain restricted map $f^{\prime}: X \xrightarrow{f} f(X)$ is a bijection and continuous by Corollary 13.2.7. Suppose $f$ is open (resp. closed) and consider any open (resp. closed) set $U \subseteq X$, then $f(U) \subseteq f(X)$ is open (resp. closed) in $f(X)$ this shows that $f^{\prime}$ is an isomorphism.

Proposition 13.2.13. A surjective embedding is an isomorphism.
Proof. Let $f: X \rightarrow Y$ be a surjective embedding. Notice that in a surjective map we have $f(X)=Y$, hence $f^{\prime}=f: X \xrightarrow{\cong} f(X)=Y$.

Example 13.2.14 (2-torus surface). Consider first a circle $(x-d)^{2}+z^{2}=r^{2}$, with centre at $(d, 0)$. Let $\phi$ be the angle going up from the $x$ to the $z$ axis, then we can parametrize such circle as $x=r \cos (\phi)+d, z=r \sin (\phi)$. Lets now consider the revolution of such circle around the $z$ axis. If $\theta$ is the angle going up from the $x$ axis to the $y$ axis, we find the new equation $\left(\sqrt{x^{2}+y^{2}}-d\right)^{2}+z^{2}=r^{2}$ has a parametrization given by

$$
(x, y, z)=((r \cos (\phi)+d) \cos (\theta),(r \cos (\phi)+d) \sin (\theta), r \sin (\phi)) .
$$

This mapping is clearly continuous. Although not injective, we can choose any point $p=(x, y)$ and find a neighbourhood of $p$ such that the mapping is injective.

Lemma 13.2.15 (Gluing). Let $X$ and $Y$ be topological spaces and consider $\mathcal{U}=\left\{U_{i}\right\}_{i}$ an open cover (or finite closed cover) of $X$. Let $\left\{f_{i}: U_{i} \rightarrow Y\right\}_{i}$ be a collection of continuous maps such that $\left.f_{i}\right|_{U_{i} \cap U_{j}}=\left.f_{j}\right|_{U_{i} \cap U_{j}}$ for all indices $i$ and $j$. Then there exists a unique continuous map $f: X \rightarrow Y$ such that $\left.f\right|_{U_{i}}=f_{i}$ for all $i$.

Proof. Let $\mathcal{U}$ be open, then given any point $p \in X,\left.f\right|_{u_{p}}=f_{p}$ is continuous for some neighbourhood $U_{p} \in \mathcal{U}$ of $p$, which implies that $f$ is continuous. On the other hand, if $\mathcal{U}$ is a finite closed cover of $X$, then consider $C \subseteq Y$ to be any closed set, then we have $f_{i}^{-1}(C)=f^{-1}(V) \cap U_{i}$ is closed in $U_{i}$ (since $f_{i}$ is continuous). Since $U_{i}$ is closed in $X$, we find that $f_{i}^{-1}(C)$ is closed in $X$ (see Proposition 13.2.3). Notice that, if $|\mathcal{U}|=n$, then $f^{-1}(C)=\bigcup_{i=1}^{n} f_{i}^{-1}(C)$ is the finite union of closed sets in $X$, which implies that $f^{-1}(C)$ itself is closed. This shows us that the inverse of $f$ maps closed sets to closed sets, which implies that $f$ is continuous. Since $\mathcal{U}$ is a cover of $X$, it is clear that $f$ is unique.

### 13.3 Product Space

Definition 13.3.1 (Product topology). Let $\left\{X_{i}\right\}_{i \in I}$ be a collection of sets. The product topology on the set $\prod_{i \in I} X_{i}$ is defined to be the initial topology such that for all $i \in I$ the projection $\pi_{i}: \prod_{i \in I} X_{i} \rightarrow X_{i}$ is continuous.

Definition 13.3.2 (Second definition of the product topology). Let $\left\{X_{i}\right\}_{i \in I}$ be a collection of topological spaces. We define the product topology on the set $\prod_{i \in I} X_{i}$ to be the topology generated by the bases

$$
\mathcal{B}=\left\{\prod_{i \in I} U_{i}: U_{i} \subseteq X_{i} \text { is open, and only finitely many } U_{i} \neq X_{i}\right\}
$$

Corollary 13.3.3. The definitions 13.3.1 and 13.3.2 are equivalent.

Proof. To prove the equivalence, we first show that if $\prod_{i \in I} X_{i}$ is endowed with the product topology, then the collection $\mathcal{B}$ is a basis for $\prod_{i \in I} X_{i} .(\Rightarrow)$ Let $\prod_{i \in I} U_{i} \in \mathcal{B}$, then for all $i_{j} \in I$ we have

$$
\pi_{i_{j}}^{-1}\left(U_{j}\right)=\prod_{i \in I} V_{i}, \text { where } V_{i}= \begin{cases}X_{i}, & i \neq i_{j} \\ U_{j}, & \text { otherwise }\end{cases}
$$

where $\pi_{i_{j}}^{-1}\left(U_{j}\right)$ is open from the hypothesis that $\prod_{i \in I} X_{i}$ is endowed with the product topology. Moreover, we also have $\prod_{i \in I} W_{i} \cap \prod_{i \in I} W_{i}^{\prime}=\prod_{i \in I} W_{i} \cap W_{i}^{\prime}$. Notice that clearly

$$
\mathcal{B}=\left\{\bigcap_{j \in J} \pi_{j}^{-1}\left(U_{j}\right): U_{j} \subseteq X_{j} \text { is open, and } J \subseteq I \text { is finite }\right\}
$$

and from Proposition 12.2.16 this shows that $\mathcal{B}$ is a basis for the product topology. ( $\Leftarrow$ ) The converse is clear.

Theorem 13.3.4 (Product topology universal product). Let $\left\{X_{i}\right\}_{i \in I}$ be a collection of topological spaces and $\pi_{i}: \prod_{i \in I} X_{i} \rightarrow X_{i}$ be the projection map. For all given topological space $Z$ and a map $f: Z \rightarrow \prod_{i \in I} X_{i}$. If $\prod_{i \in I} X_{i}$ is endowed with the product topology, then the map $f$ is continuous if and only if for every $i \in I$ the map $\pi_{i} f: Z \rightarrow \prod_{i \in I} X_{i}$. That is, the following diagram commutes in Top


Moreover, if the space $\prod_{i \in I} X_{i}$ satisfies such universal property, then its topology is the product topology.

Proof. Let $\prod_{i \in I} X_{i}$ be endowed with the product topology. $(\Rightarrow)$ Let $f$ be a continuous map and $U \subseteq Z$ be any open set. Since $\pi_{i}$ is continuous for all $i \in I$ from hypothesis then $\pi_{i}^{-1}(U)=V \subseteq \prod_{i \in I} X_{i}$ is open, we conclude that $\left(\pi_{i} f\right)^{-1}(U)=f^{-1}\left(\pi_{i}^{-1}(U)\right)=f^{-1}(V)$ is open, hence $\pi_{i} f$ is continuous. $(\Leftarrow)$ Suppose $\pi_{i} f$ is continuous, then for all given open set $U \subseteq \prod_{i \in I} X_{i}$ we have $f^{-1}\left(\pi_{i}^{-1}(U)\right) \subseteq Z$ open. Notice that since $\pi_{i}^{-1}(U)$ is open in the product topology of $\prod_{i \in I} X_{i}$ for all $i \in I$, then $f^{-1}(V) \subseteq Z$ is open for all open set $V \subseteq \prod_{i \in I} X_{i}$.

For the second part of the theorem, suppose that $\left(\prod_{i \in I} X_{i}, \tau^{\prime}\right)$ be a space satisfying the property. In particular, consider a space $Z=\left(\prod_{i \in I} X_{i}, \tau^{\prime}\right)$ and a map $f=i d$. Then the following diagram commutes


We can now assert that id is continuous since both domain and codomain have the same topology $\tau^{\prime}$, hence $\pi_{i}$ is continuous for all $i \in I$. Since $\pi_{i}$ is continuous for all $i \in I$ for the topology $\tau^{\prime}$ then we can use the Definition 13.3.1 to conclude that $\tau^{\prime} \subseteq \tau$, where $\tau$ is the product topology. For the second inclusion, consider the commutative diagram

where $\pi_{i}^{\prime}:\left(\prod_{i \in I} X_{i}, \tau^{\prime}\right) \rightarrow X_{i}$. We know from the previous discussion that since ( $\prod_{i \in I} X_{i}, \tau^{\prime}$ ) satisfies the universal property, then $\pi_{i}^{\prime}$ is continuous and hence $\mathrm{id}^{\prime}$ is continuous. In particular, this implies that if $U \subseteq\left(\prod_{i \in I} X_{i}, \tau\right)$ is open, then $\mathrm{id}^{\prime-1}(U)=U \subseteq\left(\prod_{i \in I} X_{i}, \tau^{\prime}\right)$ is open. This implies in $\tau \subseteq \tau^{\prime}$. Hence we conclude that if an object satisfies the product topology universal property, then it is endowed with the product topology.

Lemma 13.3.5 (Projections are open maps under the product topology). Let $\left\{X_{i}\right\}_{i \in I}$ be a collection of topological spaces. Then the projections

$$
\pi_{j}: \prod_{i \in I} X_{i} \rightarrow X_{j}
$$

are open maps (see Definition 12.2.9) under the product topology. Moreover, such projections are not in general closed maps.

Proof. Consider the basis $\mathcal{B}$ from Definition 13.3.2, then let $U_{i} \subseteq X_{i}$ be an open set. Then notice that (see the proof of Corollary 13.3.3)

$$
\pi_{j}\left(\pi_{i}^{-1}\left(U_{i}\right)\right)= \begin{cases}U_{i}, & i=j \\ X_{j}, & i \neq j\end{cases}
$$

which are both open sets. Moreover, we have that

$$
\pi_{j}\left(\pi_{i_{1}}^{-1}\left(U_{i_{1}}\right) \cap \cdots \cap \pi_{i_{k}}^{-1}\left(U_{i_{k}}\right)\right)= \begin{cases}X_{j}, & j \neq i \\ U_{i_{\ell}}, & j=i_{\ell} \text { for some } 1 \leqslant \ell \leqslant k\end{cases}
$$

is open. In particular, this implies in

$$
\pi_{j}\left(\bigcup_{i_{1}, \ldots, i_{k} \in I} \pi^{-1}\left(U_{i_{1}}\right) \cap \cdots \cap \pi_{i_{k}}^{-1}\left(U_{i_{k}}\right)\right)=\bigcup_{i_{1}, \ldots, i_{k} \in I} \pi_{j}\left(\pi^{-1}\left(U_{i_{1}}\right) \cap \cdots \cap \pi_{i_{k}}^{-1}\left(U_{i_{k}}\right)\right)
$$

which from the last assertion is the union of open sets, hence open. This implies that $\pi_{j}$ is open (from the fact that $\mathcal{B}$ is a basis for $\prod_{i \in I} X_{i}$ ).

On the other hand, we can build a counterexample to show why the projections are not necessarily closed. Consider the product $\mathbf{R}^{2}$ and let the closed set $C:=\{(x, y) \in$ $\left.\mathbf{R}^{2}: x y=1\right\}$, which is the hyperbola on the plane. Notice that the first projection $(x, y) \stackrel{\pi_{1}}{\longmapsto} x$ is not closed since $\pi_{1}(C)=\{x \in \mathbf{R}: x \neq 0\}$.

Definition 13.3.6 (Box topology). We define yet another natural topology on the set $\prod_{i \in I} X_{i}$ : the box topology, which is generated by the basis

$$
\mathcal{B}_{\mathrm{box}}=\left\{\prod_{i \in I} U_{i}: U_{i} \subseteq X_{i} \text { is open }\right\} .
$$

For the case where $I$ is an infinite indexing set, we find that $\tau_{\text {box }} \supsetneq \tau_{\text {prod }}$ and from this we find that certainly the projections are still continuous on $\tau_{\text {box }}$, although the box topology fails to satisfy the universal product Theorem 13.3.4.

Proposition 13.3.7 (Product of morphisms). Let $\phi: X \rightarrow Z$ and $\psi: Y \rightarrow W$ be morphisms in Top. Then the product morphism $\phi \times \psi: X \times Y \rightarrow Z \times W$ is continuous with respect to the product topology.

Proof. Consider the morphisms $\phi \pi_{X}: X \times Y \rightarrow Z$ and $\psi \times \pi_{Y}: X \times Y \rightarrow W$, so that

commutes. Therefore by the universal property of the product topology we conclude that $\phi \times \psi$ is continuous.

### 13.4 Coproduct Space

Definition 13.4.1 (First definition). Let $\left\{X_{i}\right\}_{i \in I}$ be any collection of topological spaces and consider the disjoint union $\coprod_{i \in I} X_{i}=\bigcup_{i \in I} X_{i} \times\{i\}$. We define the coproduct topology on $\coprod_{i \in I} X_{i}$ via the following property. A set $U \subseteq \coprod_{i \in I} X_{i}$ is open if and only if it is of the form $U=\coprod_{i \in I} U_{i}$, where each $U_{i} \subseteq X_{i}$ is open.

Definition 13.4.2 (Coproduct topology). Let $\left\{X_{i}\right\}_{i \in I}$ be a collection of topological spaces. The coproduct topology on the set $\coprod_{i \in I} X_{i}$ is defined to be the final topology such that for all $j \in I$ we have that the inclusions $\iota_{j}: X_{j} \mapsto \coprod_{i \in I} X_{i}$ are continuous.

Theorem 13.4.3 (Coproduct topology universal property). Let $\left\{X_{i}\right\}_{i \in I}$ be a collection of topological spaces and let $\iota_{j}: X_{j} \rightarrow \coprod_{i \in I} X_{i}$ be the $j$ th inclusion. The coproduct topology on $\coprod_{i \in I} X_{i}$ satisfies the following property. Let $Z$ be a topological space and a collection of continuous maps $\left\{f_{i}: X_{i} \rightarrow Z\right\}_{i \in I}$. Then there exists a unique map $f: \coprod_{i \in I} X_{i} \rightarrow Z$ such that the following diagram commutes in the category Top for all
$j \in I:$


On the other hand, if $\left(\coprod_{i \in I} X_{i}, \tau^{\prime}\right)$ satisfies such property, then $\tau^{\prime}$ is the coproduct topology.

Proof. (Uniqueness) Suppose that $f \iota_{j}=f_{j}$, then from definition $f_{j}(x)=f(\iota(x))=f(x, j)$ for each $j \in I$, which is clearly unique. (Existence) Suppose now that $f_{j}$ is continuous for all $j \in I$. Let $U \subseteq Z$ be any open set. Notice that since $f_{j}^{-1}=\iota_{j}^{-1} f^{-1}$ then $f_{j}^{-1}(U)=\iota_{j}^{-1}\left(f^{-1}(U)\right)$. Now, if $f^{-1}(U) \subseteq \coprod_{i \in I} X_{i}$ is closed then its preimage under $\iota_{j}$ would be closed (from the continuity of $\iota_{j}$ ), hence $f^{-1}(U)$ is open, which implies in $f$ continuous.

Suppose that $\left(\coprod_{i \in I} X_{i}, \tau^{\prime}\right)$ satisfies the universal property and denote by $\tau$ the coproduct topology. Then in particular we have

and hence $f$ and $g$ are both identities on End $\left(\coprod_{i \in I} X_{i}\right)$. If $U \subseteq\left(\coprod_{i \in I} X_{i}, \tau\right)$ is open, then $g^{-1}(U)=U \subseteq\left(\coprod_{i \in I} X_{i}, \tau^{\prime}\right)$ is also open. On the other hand, from the second diagram, if $V \subseteq\left(\coprod_{i \in I} X_{i}, \tau^{\prime}\right)$ is open, then $f^{-1}(V)=V \subseteq\left(\coprod_{i \in I} X_{i}, \tau\right)$ is open. This concludes that $\tau^{\prime}=\tau$ and hence the coproduct topology is unique.

Proposition 13.4.4. Let $\left\{X_{i}\right\}_{i \in I}$ be a collection of topological spaces. The following are properties of the coproduct space.
(a) A subset $C \subseteq \coprod_{i \in I} X_{i}$ is closed if and only if for all $X_{i}$, we have that $C \cap X_{i}$ is closed.
(b) The canonical injection $\iota_{j}: X_{j} \rightarrow \coprod_{i \in I} X_{i}$ is a topological embedding and an open and closed map.
(c) If $X_{i}$ is Hausdorff for all $i \in I$, then $\coprod_{i \in I} X_{i}$ is Hausdorff.
(d) If $X_{i}$ is first countable for all $i \in I$, then $\coprod_{i \in I} X_{i}$ is first countable.
(e) If $X_{i}$ is second countable for all $i \in I$ and $I$ is a countable indexing set, then $\coprod_{i \in I} X_{i}$ is second countable.

Properties for the coproduct top

### 13.5 Quotient Space

## Quotient Topology

The motivation for the construction of the quotient topology is the study of surjective set-functions $\pi: X \rightarrow S$ between topological spaces $X$ and sets $S$, which induce an equivalence relation on the initial topological space by means of arranging the points of $X$ in classes where $x \sim y$ for $x, y \in X$ if and only if $\pi(x)=\pi(y)$, that is, $x$ and $y$ are common points of a fibre $\pi^{-1}(s)$ for some $s \in S$.

The geometrically appealing version of such construction would be the idea of gluing every element $x \in \pi^{-1}(s)$ into a unique point - this way we loose some initial information about the topological space $X$. However, we would like to make this gluing process compatible with the theory of topology so far constructed. To achieve that, one may make the point that we should make $S$ into a topological space for which its topology makes $\pi$ a continuous map. The first idea that may come to mind is that we can simply force the continuity of $\pi$ by defining a topology $\tau$ on $S$ so that $U \subseteq S$ is an open set if and only if $\pi^{-1}(U)$ is open - which is exactly the definition of a continuous map. Lets give this a formal definition.

Definition 13.5.1. Let $X$ be a topological space and $S$ be any set. Let also $\pi: X \rightarrow S$ be a set-function. The quotient topology on $S$ induced by the map $\pi$ is the final topology such that $\pi$ is a continuous map.

Proposition 13.5.2. The quotient topology is a topology.

Proof. Let $X / \sim$ be a space with the topology induced by the map $\pi: X \rightarrow X / \sim$. Consider an arbitrary collection $\left\{U_{j}\right\}_{j}$ of open sets of $X / \sim$. Notice that since $\pi^{-1}\left(\bigcup_{j} U_{j}\right)=$ $\bigcup_{j} \pi^{-1}\left(U_{j}\right)$ then since $\pi^{-1}\left(U_{j}\right)$ is open for all index $j$, we conclude that $\bigcup_{j} \pi^{-1}\left(U_{j}\right) \subseteq X$ is open, hence the set $\bigcup_{j} U_{j} \subseteq X / \sim$ is necessarily open - since $\pi$ is continuous. Let now $A, B \subseteq X / \sim$ be any open sets, then $\pi^{-1}(A \cap B)=\pi^{-1}(A) \cap \pi^{-1}(B) \subseteq X$, which is open and therefore $A \cap B$ is open in $X / \sim$. We conclude that the quotient topology indeed satisfies the properties of a topology.

In other terms, let $T_{S}$ be the collection of all topologies on $S$ such that $\pi$ is a continuous map. The above definition simply says that the quotient topology $\tau$ on the set $S$ is the intersection $\tau=\bigcap_{T \in T_{S}} T$. Even better than that definition is the fact that we can determine the quotient topology by the following universal property.

Theorem 13.5.3 (Universal property of the quotient topology). Let $X$ be a topological space and $S$ be any set, together with a surjective set-function $\pi: X \rightarrow S$. The quotient topology $\tau$ on $S$ induced by $\pi$ is such that, for all topological spaces $Z$ and every morphism $g: X \rightarrow Z$ for which all $x, y \in X$ such that $\pi(x)=\pi(y)$ then $g(x)=g(y)$ i.e. $g$ is constant on the fibres of $\pi$ - there exists a unique morphism $f:(S, \tau) \rightarrow Z$
such that the diagram

commutes in Top. Moreover, if $\tau^{\prime}$ is a topology on $S$ such that the diagram commutes, then necessarily $\tau^{\prime}=\tau$.

Proof. Since $\pi$ is surjective, we can completely define a map $f: S \rightarrow Z$ sending $s \mapsto g(x)$ such that $x \in \pi^{-1}(s)$, which is well defined because, for all $x, y \in \pi^{-1}(s)$, we have $g(x)=g(y)$, so that the image of each $s \in S$ under the map $f$ is uniquely defined. We now show that $f$ is, in fact, continuous - and hence a morphism. Let $U \subseteq Z$ be any open set of $Z$, then, since $g^{-1}(U)$ is open and $(f \pi)^{-1}(U)=\pi^{-1} f^{-1}(U)=g^{-1}(U), f^{-1}(U)$ cannot be a closed set - in fact, it needs to be an open set, because $\pi^{-1}\left(f^{-1}(U)\right) \subseteq X$ must be open, since $\pi$ is continuous. It follows that $f$ is continuous and thus the said morphism indeed exists.

For the uniqueness, let $f$ and $f^{\prime}$ be two morphisms such that $f \pi=g$ and $f^{\prime} \pi=g$. Let $s \in S$ be any point. Since $\pi$ is surjective, there exists $x \in X$ for which $x \in \pi^{-1}(s)$, therefore, $f \pi(x)=f(s)=f^{\prime} \pi(x)=f^{\prime}(s)$ for every element of their domain - hence $f=f^{\prime}$.

Suppose now that both $(S, \tau)$ and $\left(S, \tau^{\prime}\right)$ satisfy the universal property, that is, the following diagrams commutes for unique morphisms $f$ and $h$


Since the diagram commutes for all $Z$, let $Z=(S, \tau)$, and consider the map $f^{\prime}=$ $\mathrm{id}^{\prime}:\left(S, \tau^{\prime}\right) \rightarrow(S, \tau)$. Then, given any $U \in \tau$ we find that since $\mathrm{id}^{\prime}$ is continuous that $\mathrm{id}^{\prime-1}(U)=U \subseteq\left(S, \tau^{\prime}\right)$ is open, hence $\tau \subseteq \tau^{\prime}$. Analogously, let $Z=\left(S, \tau^{\prime}\right)$ and consider $f=\mathrm{id}:(S, \tau) \rightarrow\left(S, \tau^{\prime}\right)$. Let $U^{\prime} \in \tau^{\prime}$ then from the continuity of id we find that $\mathrm{id}^{-1}\left(U^{\prime}\right)=U^{\prime} \subseteq(S, \tau)$ is open, therefore $\tau^{\prime} \subseteq \tau$. Thus indeed $\tau=\tau^{\prime}$ as wanted.

Proposition 13.5.4 (Quotient topology as a coequalizer). Let $X$ be a topological space and $f: X \rightarrow S$ be a surjective map, where $S$ is some set. Define the equivalence relation set

$$
R:=\{(x, y) \in X \times X: f(x)=f(y)\}
$$

together with two maps $r_{1}, r_{2}: R \rightrightarrows X$ given by the following commutative diagram


The quotient topology is exactly the topology that makes $S$ into the coequalizer of $r_{1}$ and $r_{2}$.

Proof. Let $Y$ be any space and $h: X \rightarrow Y$ be a continuous map such that $h r_{1}=h r_{2}$. Let $x, x^{\prime} \in X$ be any two points such that $f(x)=f(y)$, then from construction $x, x^{\prime} \in R$. Then $h(x)=h r_{1}\left(x, x^{\prime}\right)=h r_{2}\left(x, x^{\prime}\right)=h\left(x^{\prime}\right)$. Now by means of the universal property Theorem 13.5.3, if $\tau$ is the quotient topology on $S$ induced by $f$, we conclude that there exists a unique continuous map $g:(S, \tau) \rightarrow Y$ such that the following diagram commutes


Therefore $(S, \tau)=\operatorname{coeq}\left(r_{1}, r_{2}\right)$, with associated morphism $f$.
To ease the way in which we refer to quotients and surjective morphisms that induce quotients between topological spaces, we define the following terminology.

Definition 13.5.5 (Quotient morphism). A surjective morphism $\pi$ : $X \rightarrow Y$ of topological spaces $X$ and $Y$ is said to be a quotient morphism (or quotient map) if $\pi$ induces the universal property of quotients - in other words, open sets of $Y$ are exactly those that have open preimage on $\pi$, that is, $V \subseteq Y$ is open if and only if $\pi^{-1}(V) \subseteq X$ is open.

Theorem 13.5.6 (Quotient descent). Let $q: X \rightarrow Y$ be a quotient map between topological spaces. For any space $Z$ together with a morphism $f: X \rightarrow Z$ that is constant on the fibres of $q$ - that is, $q\left(x_{1}\right)=q\left(x_{2}\right)$ implies $f\left(x_{1}\right)=f\left(x_{2}\right)$ - there exists a unique morphism $f_{*}: Y \rightarrow Z$ such that the following diagram commutes


Proof. Since $q$ is surjective, for every $y \in Y$ there exists $x \in X$ such that $q(x)=y$, hence we define $f_{*}(y):=f_{*}(x)$ for every $x \in q^{-1}(y)$. For the uniqueness, since $f$ is constant on the fibres of $q$ then $f_{*}$ is fully defined by $f$ - thus unique. From the universal property we obtain that $f_{*}$ is continuous.

## Some Examples And Applications

Many important spaces can be obtained with the inclusion of the quotient topology to our toolkit, I'll now briefly discuss some of those, which will most probably come up further into this text.

Example 13.5.7 (Projective space). Let $\sim$ be the equivalence relation for which $x \sim y$ if and only if $x=\gamma y$, where $x, y \in \mathbf{R}^{n+1}\{0\}$ and $\gamma \in \mathbf{R}$. We define the $n$-dimensional real projective space as the quotient $\left(\mathbf{R}^{n+1} \backslash\{0\}\right) / \sim$, which is denoted by $\mathbf{R} \mathbf{P}^{n}$.

Definition 13.5.8 (Cone). Let $X$ be any topological space and $I$ be the standard interval. The topological space $X \times I$ is known as the cylinder on $X$. Via a quotient operation, we can collapse regions of this cylinder. For instance, we can create a cone by collapsing one of the sides of the cylinder, such as $(X \times I) /(X \times\{1\})$. Such object is denoted Cone $X$, the cone on $X$.

Definition 13.5.9 (Suspension). The suspension of a topological space is an endofunctor S: Top $\rightarrow$ Top mapping each space $X$ to the quotient space

$$
\mathrm{S} X:=(X \times I) /(X \times\{0,1\})
$$

and for each morphism $f: X \rightarrow Y$ we have the naturally induced morphism

$$
S f: S X \rightarrow S Y \text { mapping }[x, t] \mapsto[f(x), t]
$$

Example 13.5.10. Let $J$ be an indexing set and $\left\{X_{j}\right\}_{j \in J}$ be a collection of non-empty topological spaces. For each $j \in J$, choose any $p_{j} \in X_{j}$ as a base point. We define the wedge sum of the collection $\left\{X_{j}\right\}_{j \in J}$ with respect to the base points $\left\{p_{j}\right\}_{j \in J}$ as the topological space

$$
\bigvee_{j \in J} X_{j}=\coprod_{j \in J} X_{j} /\left\{p_{j}\right\}_{j \in J}
$$

An interesting fact about wedge sums of topological spaces preserve the Hausdorff property if every component is Hausdorff.

Proposition 13.5.11. Let $\left\{X_{j}\right\}_{j \in J}$ be an indexed collection of Hausdorff topological spaces. Then the wedge sum $\bigvee_{j \in J} X_{j}$ with respect to any choice of base points is Hausdorff.

Proof. Let $\left\{p_{j} \in X_{j}\right\}_{j \in J}$ be any choice of base points for the given collection. Let $x, y \in \bigvee_{j \in J} X_{j}$ be any distinct points in the wedge sum space. If $x, y \in X_{j}$ for some $j \in J$, then it is clear that there exists non-intersecting neighbourhoods of $x$ and $y$ on $X_{j}$ - of which we can take their intersection with the disjoint union $\coprod_{j \in J} X_{j}$ and the proposition will hold for $\bigvee_{j \in J} X_{j}$. On the other hand, if $i, j \in J$ are distinct indices and $x \in X_{i}$ while $y \in X_{j}$, then there exists $U_{x} \subseteq X_{i}$ and $U_{y} \subseteq X_{j}$ neighbours of $x$ and $y$, respectively, such that $U_{x} \cap X_{j}=\varnothing$ and $U_{y} \cap X_{i}=\varnothing$, which in particular imply in $U_{x} \cap U_{y}=\varnothing$. For our end, we just need to consider the neighbourhoods $U_{x}^{\prime}=U_{x} \cap \coprod_{j \in J} X_{j}$ and $U_{y}^{\prime}=U_{y} \cap \coprod_{j \in J} X_{j}$ so that $U_{x}^{\prime}=U_{y}^{\prime}$.

Proposition 13.5.12. Let $X$ be a second countable space and $M=X / \sim$ be a quotient. If $M$ is locally Euclidean, then $M$ is second countable.

Proof. Let $\pi: X \rightarrow M$ be the quotient map that induces the equivalence relation $\sim$ in $X$. If we assume that $M$ is locally euclidean, we can let $C$ be a cover of $M$ composed of coordinate balls. Since $\pi$ is surjective, $\mathcal{H}:=\left\{\pi^{-1}(U): U \in C\right\}$ is a cover for the space $X$. Moreover, since $X$ is second countable, any cover of $X$ contains a countable subcover - in particular, let $\mathcal{U} \subseteq \mathcal{H}$ be a countable subcover. Define the countable set
$C^{\prime}=\left\{U \in C: \pi^{-1}(U) \in \mathcal{U}\right\}$. Since $\mathcal{U}$ covers $X$ and $\pi$ is surjective, it follows that $C^{\prime} \subseteq C$ is a countable subcover of $M$ composed of coordinate balls - that is, $M$ is Lindelöf. Better than that, since coordinate balls are second countable (see Proposition 12.7.6) we can apply Corollary 12.5 .12 to see that $M$ is second countable.

Corollary 13.5.13 (Manifold from a quotient). In the context of the preceding proposition, if $M$ is both locally Euclidean and Hausdorff, then $M$ is a topological manifold.

## Hausdorffness \& Quotient Spaces

Remark 13.5.14. The quotient topology doesn't preserve Hausdorffness.
As an example, take the space $\mathbf{R} \times\{-1,1\}$ with the usual topology and define the equivalence relation $(x, 1) \sim(x,-1)$ if and only if $x \neq 0$. The resulting space $(\mathbf{R} \times\{-1,1\}) / \sim$ is not Hausdorff: consider the sequence $(1 / n, 1)_{n \in \mathbf{Z}_{>0}}$, we have that $1 / n \rightarrow 0$ but since $1 / n$ does never assume the value of zero, for all $n \in Z_{>0}$ we have $(1 / n, 1) \sim(1 / n,-1)$ - thus the has two distinct limits $(0,1)$ from one side and $(0,-1)$ from the other, thus $(\mathbf{R} \times\{-1,1\}) / \sim$ isn't Hausdorff.

Proposition 13.5.15 (Hausdorff from open quotients). Let $\pi: X \rightarrow Y$ be a surjective morphism of topological spaces $X$ and $Y$. Then $Y$ is Hausdorff if and only if the collection of pairs of points with common fibre, $C:=\{(p, q) \in X \times X: \pi(p)=\pi(q)\}$, is closed in $X \times X$.

Proof. Let $Y$ be Hausdorff, then, given any $(p, q) \in X \backslash C$, there are neighbourhoods $V_{p}, V_{q} \subseteq Y$ of $\pi(p)$ and $\pi(q)$, respectively, such that $V_{p} \cap V_{q}=\varnothing$. Since these neighbourhoods are disjoint, then in particular the collection fibres $\pi^{-1}\left(V_{p}\right) \times \pi^{-1}\left(V_{q}\right)$ is contained in $X \times X \backslash C$, that is, $X \times X \backslash C$ is open - hence $C$ is closed.

Let $C$ be closed, then given any distinct points $a, b \in Y$, the surjectivity of $\pi$ implies that there exists $p, q \in X$ such that $\pi(p)=a$ and $\pi(q)=b$ - in particular $(p, q) \in X \times X \backslash C$ and since $C$ is closed, there exists a neighbourhood $U_{p} \times U_{q} \subseteq X \times X$ of $(p, q)$ such that $U_{p} \times U_{q} \subseteq X \times X \backslash C$, that is, $\pi\left(U_{p}\right), \pi\left(U_{q}\right) \subseteq Y$ are non intersecting open sets (from the fact that $\pi$ is open) that are neighbourhoods of $a$ and $b$, respectively - thus $Y$ is Hausdorff.

Proposition 13.5.16. Every topological space is the quotient of a Hausdorff space.
Proof. Let $X$ be any topological space. Consider the product of real lines $P:=\prod_{x \in X} \mathbf{R}$ under the product topology and let $Y$ be the subspace of $P$ given by all points with one, and only one, rational coordinate - such space $Y$ is therefore Hausdorff. Furthermore, define a collection $\left\{Y_{x}\right\}_{x \in X}$ to consist of subsets $Y_{x} \subseteq Y$ given by the set of points whose rational coordinate has index $x$ - hence $Y_{x}$ is dense in $Y$. It should be noted that the above construction is not necessarily unique, so the reader may try to construct a different Hausdorff space $Y$ and a collection of dense sets $\left\{Y_{x}\right\}_{x \in X}$ in $Y$.

Define now the subspace of $X \times Y$ given by $Z:=\bigcup_{x \in X} x \times Y_{x}$ - endowed with the subspace topology. Let $\sim$ be the equivalence relation on $Z$ given by $\left(x_{1}, y_{1}\right) \sim\left(x_{2}, y_{2}\right)$ if and only if $x_{1}=x_{2}=x$ for some $x \in X$ and $y_{1}, y_{2} \in Y_{x}$. Define a map $\phi: Z / \sim \rightarrow X$
sending $x \times Y_{x} \mapsto x$ — which is clearly both surjective and injective. We now show that $\phi$ is a topological isomorphism.

Let $V \subseteq X$ be any open set, then $\phi^{-1}(V)=\left\{x \times Y_{x}: x \in V\right\}$ - which in turn is open in $Z / \sim$. On the other hand, let $U:=\left\{x \times Y_{x}: x \in U^{\prime}\right\}$ be any open set in $Z / \sim$ given by some indexing set of points $U^{\prime} \subseteq X-$ our goal will be to prove that $U^{\prime}$ is open, so that the inverse of $\phi$ is continuous. Consider $x_{0} \in U^{\prime}$ to be any point and likewise $\left(x_{0}, y_{0}\right) \in x_{0} \times Y_{x_{0}}$. By the fact that $\bigcup_{x \in U^{\prime}} x \times Y_{x} \subseteq Z$ is open, we are able to find neighbourhoods $O_{X}\left(x_{0}\right) \subseteq X$ and $O_{Y}\left(y_{0}\right) \subseteq Y$ - of $x_{0}$ and $y_{0}$, respectively - such that $N:=\mathrm{Z} \cap\left(O_{X}\left(x_{0}\right) \times O_{Y}\left(y_{0}\right)\right)$ is a neighbourhood of $\left(x_{0}, y_{0}\right)$ in $Z / \sim$. If $x^{\prime} \in O_{X}\left(x_{0}\right)$ is any point, then by the fact that $Y_{x^{\prime}}$ is dense in $Y$, by Proposition 12.1.34, the sets $x^{\prime} \times Y_{x^{\prime}}$ and $N$ have a non-empty intersection and thus $x^{\prime} \times Y_{x^{\prime}}$ also intersects $\bigcup_{x \in U^{\prime}} x \times Y_{x}$. Therefore $x^{\prime} \times Y_{x^{\prime}} \in U$ and hence $x^{\prime} \in U^{\prime}$ - which implies that $U^{\prime}$ is open.

## Quotient Morphisms In More Depth

So far we've been studying the construction of quotients out of surjective set-functions, but what about being able to classifying a surjective morphisms between given topological spaces as inducing the universal property of the quotient space? This will be our goal with this subsection - identifying quotient morphisms. For that end, we shall profit from the main idea behind quotients: fibres. For that, we define a set given by fibres of $f$ as being saturated.

Definition 13.5.17 (Saturated set). Let $f: X \rightarrow Y$ be a set-function. We say that a set $U \subseteq X$ is saturated with respect to $f$ if there exists $V \subseteq Y$ such that $U=f^{-1}(V)$.

Proposition 13.5.18 (Equivalences for saturated sets). Let $f: X \rightarrow Y$ be a set-function and $U \subseteq X$ be any subset. The following propositions are equivalent
(a) The set $U$ is saturated with respect to $f$.
(b) $U=f^{-1}(f(U))$.
(c) Let $p \in U$ be any point, $U$ contains every element $x \in X$ with common fibre to $p$ - that is, $f(x)=f(p)$ implies $x \in U$.

Proof. (c) $\Rightarrow$ (b): Suppose $U$ satisfies proposition (c), it is clear that $U \subseteq f^{-1}(f(U))$, on the other hand, given $x \in f^{-1}(f(U))$, it follows that $x$ has a common fibre with some point of $U$, which implies that $x \in U$. (b) $\Rightarrow$ (a): Trivial from the definition. (a) $\Rightarrow$ (c): Let $V \subseteq Y$ be such that $U=f^{-1}(V)$, then, given any $p \in f^{-1}(V)$, it is clear that $f(p) \in V$, hence every point $x \in X$ such that $f(x)=f(p) \in V$ then $x \in U$, which finishes the equivalence chain.

Proposition 13.5.19 (Classification of surjective morphisms). Let $\pi: X \rightarrow Y$ be a surjective morphism of topological spaces. The map $\pi$ is a quotient morphism - that is, induces the universal property of quotients for $X$ and $Y$ - if and only if every saturated open (or closed) set of $X$ has an open (or closed) image in $Y$.

Proof. Let $\pi$ be any surjective morphism taking saturated open sets to open images. Let $V \subseteq Y$ be an open set. Since $\pi$ is surjective and continuous, then $\pi^{-1}(V) \subseteq X$ is open. On the other hand, let $V \subseteq Y$ be any set of $Y$ (not necessarily open), such that $\pi^{-1}(V):=U \subseteq X$ is open. This implies directly that $U$ is saturated with respect to $\pi$ and from our initial hypothesis, $\pi(U)=V \subseteq Y$ is open. Thus $\pi$ is a quotient morphism.

For the contrary, let $\pi$ be a quotient morphism. Then, given any $U \subseteq X$ open set, saturated with respect to $\pi$, define $V \subseteq Y$ such that $U=\pi^{-1}(V)$. Since $\pi$ is a quotient morphism, it follows that $V$ is necessarily open in $Y$, thus $\pi(U)=V$ is open.

The proof for the closed set case is completely analogous.
■

Proposition 13.5.20 (Properties of quotient morphisms). The following properties pertain to quotient morphisms between topological spaces.
(a) The composition of quotient morphisms is a quotient morphism.
(b) Injective quotient morphisms are isomorphisms.
(c) Let $\pi: X \rightarrow Y$ be a quotient morphism. Then, $C \subseteq Y$ is closed if and only if $\pi^{-1}(C) \subseteq X$ is closed.
(d) Let $\pi: X \rightarrow Y$ be a quotient morphism and $U \subseteq X$ be any saturated set (open or closed) with respect to $\pi$. Then, the restriction $\left.\pi\right|_{U}: U \rightarrow \pi(U)$ is a quotient map.
(e) Let $J$ be an indexing set and $\left\{\pi_{j}: X_{j} \rightarrow Y_{j}\right\}_{j \in J}$ be an indexed collection of quotient morphisms. The map $\pi: \coprod_{j \in J} X_{j} \rightarrow \coprod_{j \in J} Y_{j}$ defined by the restrictions $\pi\left(x_{j}\right)=$ $\pi_{j}\left(x_{j}\right)$ for every $x_{j} \in X_{j} \cap \coprod_{j \in J} X_{j}$ is a quotient map.

Proof. (a) Let $\pi: X \rightarrow Y$ and $\pi^{\prime}: Y \rightarrow Z$ be quotient morphisms, and consider the map $\pi^{\prime} \pi: X \rightarrow Z$, which is clearly surjective. From hypothesis a subset $U \subseteq Z$ is open if and only if $\pi^{\prime-1}(U)$ is open, moreover, $\pi^{\prime-1}(U) \subseteq Y$ is open if and only if $\pi^{-1}\left(\pi^{\prime-1}(U)\right) \subseteq X$ is open - the proposition follows.
(b) If the quotient morphism $\pi: X \rightarrow Y$ is injective, then $\pi$ is a bijection. Let $U \subseteq X$ be any open set, since $\pi$ is bijective, there exists a set $V \subseteq Y$ such that $\pi^{-1}(V)=U$ - moreover, such set must be open on $Y$ from the quotient topology. This shows that $\pi(U)=V$ is open and hence $\pi$ is a topological isomorphism.
(c) Let $C \subseteq Y$ be any set, notice that $Y \backslash C$ is open if and only if $\pi^{-1}(Y \backslash C)=$ $\pi^{-1}(Y) \backslash \pi^{-1}(C) \subseteq X$ is open - thus the proposition follows.
(d) Let $V \subseteq \pi(U)$ be any set. Since $\pi$ is a quotient morphism, $V$ is open if and only if $\pi^{-1}(V) \subseteq U \subseteq X-$ moreover, since $U$ is saturated, the whole set can have its subsets classified by $\left.\pi\right|_{U}$ into open or closed sets, thus $\left.\pi\right|_{U}$ is a quotient map.
(e) Let $U \subseteq \coprod_{j \in J} Y_{j}$ be any set and let $j_{0} \in J$ be such that $U \subseteq Y_{j_{0}}$. From the mappings, $U$ is open if and only if $\pi_{j_{0}}^{-1}(U) \subseteq X_{j_{0}}$ is open — but since $\pi^{-1}(U)=\pi_{j_{0}}^{-1}(U)$, the proposition follows.

Example 13.5.21 (Cones). An application of the last proposition takes us back to Definition 13.5.8, where we defined the cone Cone $X$ of a topological space $X$ as
$(X \times I) /(X \times\{1\})$ - the collapse the top of the cylinder. Notice that, given any point $(x, t) \in(X \times I) \backslash(X \times\{1\})$, there exists a neighbourhood $U \subseteq(X \times I) \backslash(X \times\{1\})$ of $(x, t)$, therefore $X \times\{1\}$ is closed in $X \times I$. If we consider the quotient morphism of the cone $\pi: X \times I \rightarrow$ Cone $X$, the restriction $\left.\pi\right|_{X \times\{0\}}: X \times\{0\} \rightarrow \pi(X \times\{0\})$ is also a quotient map - moreover, such quotient map is injective, thus $\left.\pi\right|_{X \times\{0\}}$ is an isomorphism. Therefore we have a sequence of isomorphisms

$$
X \xrightarrow{\simeq} X \times\{0\} \xrightarrow{\simeq} \pi(X \times\{0\}) \subseteq \text { Cone } X
$$

thus we can identify $X$ as a subspace of Cone $X$.

The following is a sufficient, but not necessary condition for a quotient morphism.

Proposition 13.5.22. Let $\pi: X \rightarrow Y$ be a surjective topological morphism. If $\pi$ is either open or closed, then $\pi$ is a quotient morphism.

Proof. If $\pi$ is open (respectively, closed), in particular we have that saturated sets open (respectively, closed) of $X$ are mapped to open (respectively, closed) sets of $Y$, hence $\pi$ is a quotient morphism.

Proposition 13.5.23. Let $f: X \rightarrow Y$ be a topological morphism that is either open or closed. The following are properties hold:
(a) If $f$ is injective, it is a topological embedding.
(b) If $f$ is surjective, it is a quotient map.
(c) If $f$ is bijective, it is an isomorphism.

Proof. We work out the proof for the case where $f$ is open, the closed case is equivalent:

1. Consider the restriction of the codomain $f^{\prime}: X \rightarrow f(X)$ - which is certainly surjective, thus bijective. Let $g: f(X) \rightarrow X$ denote the inverse of $f$. Notice that since $f$ is open then any $U \subseteq X$ open implies in $f(U) \subseteq Y$ also open, therefore $g(U)=f(U)$ is open and hence $g$ is continuous.
2. Let $U \subseteq Y$ be any set. Since $f$ is surjective, there exists $V \subseteq X$ such that $f(V)=U$. Moreover, since $f$ is open, $U$ can only be open in $Y$ if its preimage $f^{-1}(U)=V$ is open in $X$ - which implies that $f$ is a quotient map.
3. If $f$ is bijective, then by the fist item $f$ is an embedding, that is, it yields an isomorphism $X \simeq f(X)$. Moreover since $f$ is also surjective then $f(X)=Y$ and thus $f$ is an isomorphism $X \xrightarrow{\simeq} Y$.

### 13.6 Attaching Space

Definition 13.6.1 (Attaching space). Let $X$ and $Y$ be topological spaces, and consider a subspace $A \subseteq X$ together with a continuous map $f: A \rightarrow X$. The attaching space of $X$ and $Y$ along $f$ is defined to be the pushout of the canonical inclusion $t: A \hookrightarrow X$ and $f$, that is


Corollary 13.6.2. In the notation of Definition 13.6.1, the attaching space of $X$ and $Y$ along $f$ is given by the quotient space

$$
X \cup_{f} Y \simeq(X \amalg Y) / \sim,
$$

where $\sim$ is the smallest equivalence relation on $X \amalg Y$ such that $x \sim f(x)$ for all $x \in A$.
Proof. Let $Z$ be any space together with two continuous maps $p: X \rightarrow Z$ and $q: Y \rightarrow Z$ such that $p \iota=q f$ - that is, for every $x \in A$ we have $p(x)=q(f(x))$. We define a $\operatorname{map} \phi:(X \amalg Y) / \sim \rightarrow Z$ by $[(x, X)] \mapsto p(x)$ and $[(y, Y)] \mapsto q(y)$. Indeed, the image of a class point under $\phi$ does not depend on the representative since, for any $x \in A$, we have $[(x, X)] \sim[(f(x), Y)]$, but $p(x)=q(f(x))$. Also, since $[X]:=\{[(x, X)]\}_{x \in X}$ and $[Y]:=\{[(y, Y)]\}_{y \in Y}$ are open subspaces covering $(X \amalg Y) / \sim$, since $\left.\phi\right|_{[X]}=p$ and $\left.\phi\right|_{[Y]}=q$ are continuous maps, by Proposition 13.2 .9 we conclude that $\phi$ is continuous.

If we consider the inclusions $\iota_{X}: X \hookrightarrow(X \amalg Y) / \sim$ and $\iota_{Y}: Y \hookrightarrow(X \amalg Y) / \sim$, one has that $\phi \iota_{X}=p$ and $\phi \iota_{Y}=q$. On the other hand, since these inclusions are monomorphisms in Top, we conclude that $\phi$ is the only continuous map making the following diagram commutative in Top


We conclude that $(X \amalg Y) / \sim$ is the pushout of $f$ with $\iota$ and since pushouts are unique up to isomorphism, then $(X \amalg Y) / \sim \simeq X \cup_{f} Y$.

Proposition 13.6.3 (Attaching space properties). Let $A \subseteq Y$ be a closed subspace, $f: A \rightarrow$ $X$ be a continuous map, and $\pi: X \amalg Y \rightarrow X \cup_{f} Y$ be the canonical projection to the attaching space along $f$. The following properties hold:
(a) The restriction $\left.\pi\right|_{X}$ is a topological embedding, and its image $\pi(X)$ is a closed subspace of $X \cup_{f} Y$.
(b) The restriction $\left.\pi\right|_{Y \backslash A}$ is a topological embedding, and its image $\pi(Y \backslash A)$ is an open subspace of $X \cup_{f} Y$.
(c) The attaching space $X \cup_{f} Y$ is isomorphic to the disjoint union $\pi(X) \amalg \pi(Y \backslash A)$.

Proof. (a)
Prove attaching space properties and continue on attaching manifolds along the boundary.

## Chapter 14

## Connectivity and Compactness

### 14.1 Connected Spaces

## Connectedness

Definition 14.1.1 (Connected space). Let $X$ be a topological space. We say that $X$ is connected if and only if one of the following conditions hold:
(a) The space $X$ cannot be expressed as the union of two disjoint non-empty open sets.
(b) Every morphism $f: X \rightarrow\{0,1\}$ is constant-where $\{0,1\}$ is a space endowed with the discrete topology.

Corollary 14.1.2. The two conditions in Definition 14.1.1 are equivalent.
Proof. Assume $X$ satisfies the condition (a), and let $x \in X$ be any point. Suppose $f(x)=0$ (or $f(x)=1$, we do not loose generality by choosing a point of the domain), and consider $f^{-1}(0) \subseteq X$, which must be open from the continuity of $f$, thus $f^{-1}(0)$ is a neighbourhood of $x$ in $X$. If we suppose, for the sake of contradiction, that there exists $y \in X$ such that $f(y)=1$, then $f^{-1}(1)$ is also open and is a neighbourhood of $y$-notice however that $f^{-1}(0) \cap f^{-1}(1)=\varnothing$, thus we arrive at a contradiction, there must be no point $y$ with image different than zero.

We prove the counter positive: not (b) implies not (a). Suppose $f: X \rightarrow\{0,1\}$ is not constant, so that there exists two points $x, y \in X$ such that $f(x)=0$ and $f(y)=1$ (and therefore $f$ is surjective), notice however that, since $f$ is continuous,

Definition 14.1.3 (Connected components). Let $X$ be a topological space. Define an equivalence relation $\sim$ by, given $x, y \in X$, we have $x \sim y$ if and only if there exists a connected subspace of $X$ containing both $x$ and $y$. The collection of equivalence classes $X / \sim$ is called connected components of $X$.

Definition 14.1.4 (Totally disconnected). A space is said to be totally disconnected if the only connected subsets are singletons.

Example 14.1.5. The set of rational numbers $\mathbf{Q}$ is totally disconnected.

## Path Connectedness

Notation 14.1.6 (Standard interval). From now on, when talking about paths and homotopies, we shall reserve the symbol $I$ to denote the standard topological interval, which is defined by

$$
I:=[0,1] \hookrightarrow \mathbf{R} .
$$

Definition 14.1.7 (Paths \& loops). A path in a topological space $X$ is any continuous map $\gamma: I \rightarrow X$. A loop in $X$ is a continuous map $\ell: I \rightarrow X$ such that $\ell(0)=\ell(1)$.

Definition 14.1.8 (Path connected space). A topological space $X$ is said to be path connected if and only if for all $x, y \in X$ there exists a path connecting $x$ and $y$.

Proposition 14.1.9 (Path connected equivalence relation). There exists an equivalence relation $\sim$ on the topological $X$ defined by: given $x, y \in X$, we have $x \sim y$ if and only if there exists a path in $X$ connecting $x$ and $y$.

Proof. The constant path $x: I \rightarrow X$ given by $t \mapsto x$ is a path on $X$, thus $x \sim x$. Let $x \sim y$ and $\gamma$ be a path from $\gamma(0)=x$ to $\gamma(1)=y$, then we can define a map $\lambda:[0,1] \rightarrow X$ given by $\lambda(t):=\gamma(1-t)$, so that $\lambda$ is both continuous, and $\lambda(0)=y$ while $\lambda(1)=x$-thus $\lambda$ is a path between $y$ and $x$, therefore $y \sim x$. Suppose now $y \sim z$ and let $\eta$ be a path connecting $y$ to $z$. We define a map $\phi: I \rightarrow X$ given by

$$
\phi(t):= \begin{cases}\lambda(2 t), & \text { for } t \in[0,1 / 2]  \tag{b}\\ \eta(2 t-1), & \text { for } t \in[1 / 2,1]\end{cases}
$$

which is surely continuous and connects both $x$ and $z$-thus $x \sim z$.
Definition 14.1.10 (Path connected components). Let $X$ be a path connected topological space, and $\sim$ be the equivalence relation described in Proposition 14.1.9. The collection $X / \sim$ is called the path connected components of the space $X$. We denote the collection of all path components of $X$ by $\pi_{0}(X)$-the collection of homotopy classes between maps * $\rightarrow$ X.

Definition 14.1.11 ( $\pi_{0}$ functor). The concept of connected components of a space induce a covariant functor $\pi_{0}$ : Top $\rightarrow$ Set defined by mapping objects $X \mapsto \pi_{0} X$, and morphisms $f: X \rightarrow Y$ to $\pi_{0} f: \pi_{0} X \rightarrow \pi_{0} Y$-which is a well defined map since, given a path component $P$ of $X$, the set $f(X) \subseteq Y$ is connected and therefore contained in a unique path component of $Y$.

## Properties of Connectivity

Theorem 14.1.12 (Morphisms preserve connectivity). Let $X$ be a (path) connected space and $f: X \rightarrow Y$ be a topological morphism. Then $f(X) \subseteq Y$ is (path) connected.

Proof. Suppose that $f(X)$ is not connected, and let $g: f(X) \rightarrow\{0,1\}$ be a non-constant morphism-in particular, this is equivalent to the condition of $g f: X \rightarrow\{0,1\}$ being a non-constant morphism, thus implying in the non-connectedness of $X$.

On the other hand, assume now that $X$ is path connected and let $u, v \in f(X)$ define $x, y \in X$ so that $f(x)=u$ and $f(y)=v$. Let $\gamma: I \rightarrow X$ be a path connecting $x$ and $y$-then $f \gamma: I \rightarrow f(X)$ is a path connecting $u$ and $v$, which proves the proposition.

Corollary 14.1.13. Connectedness and path connectedness are both topological properties.

Corollary 14.1.14. The quotient of a (path) connected topological space is (path) connected.

Proposition 14.1.15. Let $f: X \rightarrow Y$ be a morphism of topological spaces $X$ and $Y$. If $Y$ is connected and, for all $y \in Y$, the fibre $f^{-1}(y)$ is connected, then $X$ is connected.

Proof. Let $g: X \rightarrow\{0,1\}$ be a morphism. From the connectedness condition on the fibres of $f$, it follows that $g$ must be constant throughout the fibres of $f$-this implies in the existence of a morphism $g^{*}: Y \rightarrow\{0,1\}$ such that $g=g^{*} f$. Since $Y$ is connected, $g^{*}$ must be constant, therefore the composition $g^{*} f$ is constant and so is $g$-that is, $X$ is connected.

Proposition 14.1.16. Let $X$ be a topological space. The following are properties concerning connectivity:
(a) Let $U$ and $V$ be disjoint open subsets of $X$. If $A \subseteq X$ is connected and is contained in $U \cup V$, then either $A \subseteq U$ or $A \subseteq V$.
(b) If $X$ contains a dense connected set, then $X$ is connected.
(c) Let $A \subseteq X$ be a connected set. Then $\mathrm{Cl}(A)$ is connected and any subset $B \subseteq X$ with $A \subseteq B \subseteq \mathrm{Cl}(A)$ is connected.

Proof. (a) Suppose $A$ has points in both sets, then $A \cap U$ and $A \cap V$ would be two disjoint non-empty sets whose union is $A$-which would imply that $A$ is not connected.
(b) Let $D \subseteq X$ be a dense connected set and suppose there exists non-empty disjoint open sets $U$ and $V$ whose union is $X$. Therefore $D \subseteq U \cup V$ and by item (a) we have $D \subseteq U$ or $D \subseteq V$. If $D \subseteq U$, then $X=\mathrm{Cl}(D) \subseteq \mathrm{Cl}(U)$ but since $\mathrm{Cl}(U) \subseteq X$ it follows that $V=\varnothing$.
(c) Notice that since $A \subseteq B \subseteq \mathrm{Cl}(A)$ then $\mathrm{Cl}(A)=B$ and from item (b) we conclude that $B$ is connected.

Proposition 14.1.17. Let $J$ be a set, and consider a collection $\left(X_{j}\right)_{j \in J}$ of (path) connected topological spaces. Define the space $X:=\bigcup_{j \in J} X_{j}$. If the intersection $\bigcap_{j \in J} X_{j}$ is nonempty, then $X$ is (path) connected.

Proof. Suppose the intersection is non empty and let $x \in \bigcup_{j \in J} X_{j}$ be any point. We split the proposition into the two given cases.

Suppose $X_{j}$ is connected for all $j \in J$. Let $f: X \rightarrow\{0,1\}$ be a morphism-we want to show that it has to be constant. Since $X_{j}$ is connected, it follows that the restriction
$\left.f\right|_{X_{j}}$ has to be constant, and, since $x \in X_{j}$, then $f(x)=f(y)$ for all $y \in X_{j}$. The fact that this must be true for all $X_{j}$ shows that $f$ must be constant throughout the whole set $X$.

Suppose $X_{j}$ is path connected for all $j \in J$-that is, from hypothesis, $X_{j}$ is a path connected component of $X$. We now prove that the intersection is connected to each of the path connected components. For each $j \in J$, choose any point $y \in X_{j}$. Since $x \in X_{j}$, then $x \sim y$ and therefore every point of $X$ is connected by a path.

Theorem 14.1.18. In the real space $\mathbf{R}$, the only connected sets are intervals and singletons.

Proof. If $A$ is not an interval, there must exist a pair $x, y \in A$ for which there is $z \notin A$ with $x<z<y$. This way, we can write $A$ as the union of two disjoint non-empty sets: $A=[A \cap(-\infty, z)] \cup[A \cap(z, \infty)]$-that is, $A$ is disconnected.

We now show that intervals are connected. For the sake of contradiction, assume there exists an interval $I$ such that there are disjoint and non-empty sets $A$ and $B$ for which $I=A \cup B$-that is, we assume $I$ is disconnected. Let $x, y \in I$ be any two elements with $x<y$ and $x \in A$, while $y \in B$. By the archimedean principle, valid in $\mathbf{R}$, since the set $C:=[x, y) \cap A$ is both non-empty and bounded above, thus $C$ has a supremum, let $s:=\sup C-$ where $x<s \leqslant y$. Notice that $s \notin A$, since if so then $s$ would not be the supremum of $C$, therefore $s \in B$-which also cannot be the case, since $s$ would not, again, be the supremum of $C$. We conclude that the sets $A$ and $B$ cannot be constructed at all, hence there is no disconnected interval in $\mathbf{R}$.

Proposition 14.1.19 (Partition by components). The (path) connected components of any topological space form a partition of the space.

Proof. Let $X$ be a topological space. Let $U$ and $V$ be (path) connected components of $X$ and suppose that $U \cap V$ is non-empty. Since the union of (path) connected sets is (path) connected, it follows that $U \cup V$ is (path) connected from Proposition 14.1.17. Since (path) connected components are maximal, it follows that $U \cup V=U=V$. Therefore the intersection of (path) connected components is always empty.

Let $p \in X$ be any point. Considering the (path) connected set $\{p\}$, one finds a maximal (path) connected component $U$ containing $\{p\}$. Therefore the collection of (path) connected components covers $X$.

Corollary 14.1.20. Let $X$ be a non-empty topological space. The following are properties concerning the components of $X$ :
(a) Any non-empty (path) connected subset of $X$ is contained in a single connected component.
(b) Any non-empty path connected subset of $X$ is contained in a single path connected component.
(c) Each connected component of $X$ is the disjoint union of path components.

Proof. For the proof of (a), let $A \subseteq X$ be a non-empty (path) connected set. Let $p \in A$ be any point. Since connected components cover $X$, there exists a connected component
$C \subseteq X$ such that $p \in C$. Since the union of connected sets is connected, then $A \cup C$ is connected (if $A$ is path connected, then it's also connected). Since $A \cup C$ contains $C$, by the maximality of the components it must be the case that $A \cup C=C$. Therefore $A \subseteq C$. For item (b) we have a completely analogous proof.

For the proof of item (c), let $C$ be a connected component of $X$. For any point $p \in C$, since path components cover $X$, let $P$ be a path component containing $p$. Since path components are connected sets, the union $P \cup C$ is connected and contains $C$. Again, by the maximality of $C$ this implies in $P \subseteq C$. Since path components are also maximal and any point of $C$ can be found in a path component contained in $C$, it follows that $C$ is the disjoint union of path components.

Lemma 14.1.21. The connected components of a non-empty topological space are closed.

Proof. Let $C$ be a connected component of a topological space $X$. Since $C$ is dense in $\mathrm{Cl}(\mathrm{C})$ it follows that $\mathrm{Cl}(\mathrm{C})$ is connected (see Proposition 14.1.16 item (b)). Since connected components are maximal, it follows that $\mathrm{Cl}(\mathrm{C})=\mathrm{C}$, thus C is closed.

Proposition 14.1.22. Any path connected space is connected.
Proof. Let $X$ be a path connected space and $f: X \rightarrow\{0,1\}$ be any continuous map. Let $x, y \in X$ be any two points. Since $X$ is path connected, there exists a path $\gamma: I \rightarrow X$ between $x$ and $y$. From the hypothesis that $f$ is continuous, $f$ must be continuous on $\gamma(I)$, therefore constant. Notice that, since this is true for all points of $X$ we find that $f(x)=f(y)$ for all $x, y \in X$.

Proposition 14.1.23. Connectedness and path connectedness are homotopy invariants.
Proof. Let $f: X \xrightarrow{\simeq} Y$ be a homotopy equivalence, and $g: Y \xrightarrow{\simeq} X$ be its homotopy inverse-furthermore, consider a homotopy $h: f g \rightarrow \mathrm{id}_{\gamma}$.

Suppose $X$ is connected, we shall prove that $Y$ is connected. Let $k: Y \rightarrow\{0,1\}$ be any continuous map and $y, y^{\prime} \in Y$ be any two points. Since $X$ is connected, in particular the map $k f: X \rightarrow\{0,1\}$ is constant-therefore, $k f g(y)=k f g\left(y^{\prime}\right)$. Since $h$ is a homotopy, then the induced maps $h(y,-), h\left(y^{\prime},-\right): I \rightrightarrows Y$ are such that $h(y, 0)=f g(y)$ and $h(y, 1)=y$, while $h\left(y^{\prime}, 0\right)=f g\left(y^{\prime}\right)$ and $h\left(y^{\prime}, 1\right)=y^{\prime}$. We conclude that $h(y,-)$ is a path from $f g(y)$ to $y$ and $h\left(y^{\prime},-\right)$ is a path from $f g\left(y^{\prime}\right)$ to $y^{\prime}$. Since $k$ is continuous, it follows that $k f g(y)=k(y)$ and $k f g\left(y^{\prime}\right)=k\left(y^{\prime}\right)$. Finally we obtain $k(y)=k\left(y^{\prime}\right)$, showing that $k$ is constant.

Suppose $X$ is path connected. Then $f(X)$ is path connected and therefore $k$ must be constant in every point of $f(X)$. Now, if $y \in Y \backslash f(X)$, we consider the induced map $h(y,-): I \rightrightarrows Y$-which is a path from $f g(y) \in f(X)$ to $y$. Therefore, every point of $Y \backslash f(X)$ can be connected by a path to a point of $f(X)$, proving that $Y$ itself is path connected.

Proposition 14.1.24 (Products preserve connectedness). Let $\left(X_{j}\right)_{j \in J}$ be a collection of (path) connected topological spaces. Then $\prod_{j \in J} X_{j}$ is (path) connected.

Proof. Lets define the notation $X:=\prod_{j \in J} X_{j}$.
Suppose $\left(X_{j}\right)_{j \in J}$ is composed of connected spaces. Let $k: X \rightarrow\{0,1\}$ be any continuous map. For every $j_{0} \in J$, let $p \in \prod_{j \in J \backslash j_{0}} X_{j}$ be any point-which, from construction, excludes the $j_{0}$-th coordinate from the original product space $X$. Consider the continuous map $\iota_{j_{0}}: X_{j_{0}} \rightarrow X$ for which $\pi_{j} \iota_{j_{0}}(x):=\pi_{j}(p)$ for $j \neq j_{0}$ and $\pi_{j_{0}} \iota_{j_{0}}(x):=x$-that is, $\iota_{j_{0}}$ embeds $X_{j_{0}}$ in $X$ where every coordinate, but $j_{0}$, is fixed using the pre-chosen point $p$. Since $k$ is continuous and $X_{j}$ is connected, the continuous map $k \iota_{j_{0}}: X_{j} \rightarrow\{0,1\}$ has to be constant. Now, if we take any pair of points $x, y \in X$, one sees that $k(x)=k(y)$ from the fact that $k \iota_{j}$ has to be constant for all $j \in J$-for the suitable choice of initial fixed point.

Suppose $\left(X_{j}\right)_{j \in J}$ is a collection of path connected spaces. Let $x, x^{\prime} \in X$ be any two elements. For each $j \in J$, let $\gamma_{j}: I \rightarrow X$ be a path connecting $\gamma_{j}(x)$ to $\gamma_{j}\left(x^{\prime}\right)$. By the universal property of the product topology, there exists a unique continuous map $\gamma: I \rightarrow X$ such that $\pi_{j} \gamma=\gamma_{j}$ for every $j \in J$. Since the product of the paths, $\gamma$, is a path from $x$ to $x^{\prime}$, we conclude that $X$ is path connected.

Theorem 14.1.25. A space $X$ is connected if and only if the covariant functor $\operatorname{Mor}_{\text {Top }}(X,-)$ preserves coproducts.

## Prove connectedness iff preserves coproducts.

## Local Connectedness

Definition 14.1.26 (Locally (path) connected space). A space $X$ is said to be locally (path) connected if it admits a basis of (path) connected open subsets. Equivalently, for every $p \in X$ and neighbourhood $U \subseteq X$ of $p$, there exists a (path) connected neighbourhood $V \subseteq U$ of $p$.

Proposition 14.1.27 (Properties for locally (path) connected spaces). Let $X$ be a locally (path) connected space. Then the following properties hold:
(a) Every open subset of $X$ is locally (path) connected.
(b) Every (path) connected component of $X$ is open.

Proof. Since any open subset $A$ can be written as the union of elements of the basis, then the elements of the basis of $X$ contained in $A$ form a basis of (path) connected sets for $A$. Therefore $A$ is locally (path) connected.

Let $C$ be a (path) connected component of $X$, and $p \in C$ be any point. Since $X$ is locally (path) connected, there exists a (path) connected neighbourhood $U \subseteq X$ of $p$. Therefore by Corollary 14.1 .20 we find that $U \subseteq C$. Therefore $C$ is open.

Proposition 14.1.28 (Properties of locally path connected spaces). Let $X$ be a locally path connected space. Then the following properties hold:
(a) $X$ is locally connected.
(b) The path components of $X$ are equal to its connected components.
(c) $X$ is connected if and only if it is path connected.

Proof. (a) From Proposition 14.1.22 we find that the path connected basis of $X$ is also a connected basis, therefore $X$ is locally connected.
(b) Let $p \in X$ be any point and consider $P$ and $C$ to be, respectively, the path component and the connected component containing $p$. Since $P$ contains a point of $C$, then $P \subseteq C$-since $P \cup C$ is connected and contains $C$. Moreover, by Corollary 14.1.20 we know that $C$ can be written as the disjoint union of path components.
Moreover, it needs to be the case that $P$ is the only path component in $C$. On the contrary, then $P$ and $C \backslash P$ would be non-empty disjoint open sets whose union is $C$-making $C$ disconnected. Therefore $C=P$.
(c) $X$ is connected if and only if it has a unique connected component. Since path components equal connected components, it follows that uniqueness of connected components is equivalent to uniqueness of path components. Since $X$ is path connected if and only if it has a unique path component, the proposition follows.

Proposition 14.1.29 (Connectivity of manifolds). Let $M$ be an $n$-manifold with or without boundary. The following properties hold:
(a) $M$ is locally path connected.
(b) $M$ has countably many connected components, each being an open subset of $M$ and a connected topological space.

Proof. (a) Since $M$ has a basis of coordinate balls and those are isomorphic to an open subset of $\mathbf{R}^{n}$, which are path connected, it follows that the coordinate balls are path connected.
(b) By item (a) and Proposition 14.1.27, the collection of connected components of $M$ forms an open cover of $M$. Since $M$ is second countable, there exists a countable subcover. Since connected components are disjoint, the subcover must be the cover itself, showing that the collection of connected components of $M$ is countable.

Since connected components are open, together with the initial topology generated by the canonical inclusion-that is, the subspace topology-we find that connected components are topological manifolds.

Corollary 14.1.30. A connected manifold is path connected.
Proof. Follows directly from Propositions 14.1.28 and 14.1.29.

### 14.2 Compact Spaces

Definition 14.2.1 (Compact space). A topological space $X$ is said to be compact if for every open cover there exists a finite subcover.

Proposition 14.2.2 (Image of compact space). If $f: X \rightarrow Y$ is a topological morphism and $X$ is compact, then $f(X)$ is compact in $Y$.

Proof. Let $C$ be an open cover of $f(X)$. Consider the preimage collection $\mathcal{U}:=$ $\left\{f^{-1}(V)\right\}_{V \in C}$, which by construction covers $X$. Since $X$ is compact, let $\mathcal{U}^{\prime}$ be the finite subcover given by $\mathcal{U}$. If we now consider the image collection $C^{\prime}:=\{f(U)\}_{U \in C^{\prime}}$, we find that $C^{\prime}$ covers $f(X)$ and is contained in $C$, therefore a finite subcover.

Corollary 14.2.3. Compactness is an invariant property of topological spaces.
Corollary 14.2.4. The quotient of a compact space is compact.
Proof. Since any quotient space is the image of a continuous projection, the proposition follows from Proposition 14.2.2.

Proposition 14.2.5 (Closed subset is compact). In a compact space any closed subset is compact.

Proof. Let $X$ be a space and $A \subseteq X$ any closed subset. If $\mathcal{U}$ is an open cover of $A$, then $\mathcal{U} \cup\{X \backslash A\}$ is a cover of $X$. Moreover, since $X$ is compact, there exists a finite subcover $\mathcal{U}^{\prime} \subseteq \mathcal{U} \cup\{X \backslash A\}$. In particular, since $\mathcal{U}^{\prime}$ covers $X$, it also covers $A$.

## Checking if a Space is Compact

Definition 14.2.6 (Finite intersection property). A collection of sets $\mathcal{A}$ is said to satisfy the finite intersection property if and only if for every finite collection $\left\{A_{1}, \ldots, A_{n}\right\} \subseteq \mathcal{A}$, we have $\bigcap_{j=1}^{n} A_{j} \neq \varnothing$, that is, the intersection is non-empty

Proposition 14.2.7. A space is compact if and only if every collection of closed subsets satisfying the finite intersection property has non-empty intersection.

Proof. Let $X$ be a compact space and $C$ be any collection of closed subsets satisfying the finite intersection property. Suppose, for the sake of contradiction, that $C$ has an empty intersection-thus the complement of the intersection covers $X$. Hence there must exist a finite collection $\left\{A_{1}, \ldots, A_{n}\right\} \subseteq C$ whose corresponding complement covers $X$ and therefore $\bigcup_{j=1}^{n}\left(X \backslash A_{j}\right)=X \backslash\left(\bigcap_{j=1}^{n} A_{j}\right)=X$-then $\bigcap_{j=1}^{n} A_{j}=\varnothing$, which contradicts the hypothesis that $C$ satisfies the finite intersection property.

Suppose that the latter condition is true. Suppose there exists an open cover $\mathcal{U}$ of $X$ that has no finite subcover. In particular, the collection $\mathcal{U}^{\prime}:=\{X \backslash U: U \in$ $\mathcal{U}\}$ is composed of closed sets and satisfies the finite intersection property-thus $\cup_{U \in \mathcal{U}}(X \backslash U)=X \backslash\left(\bigcap_{U \in \mathcal{U}} U\right)=X$, which immediately implies that the intersection of $\mathcal{U}$ is empty, yielding a contradiction. Hence $\mathcal{U}$ must have a finite subcover.

Lemma 14.2.8 (Non-empty countable intersection). Let $X$ be a compact space and $\left\{C_{j}\right\}_{j \in \mathbf{N}}$ be a countable collection of non-empty subsets of $X$ for which $C_{j+1} \subseteq C_{j}$ for every $j \in \mathbf{N}$-that is, the sets are nested. Then the countable intersection $\bigcap_{j \in \mathbf{N}} C_{j}$ is non-empty.

Definition 14.2.9 (Proper map). A continuous map $f: X \rightarrow Y$ is said to be proper if for all compact sets $C \subseteq Y$ the preimage $f^{-1}(C) \subseteq X$ is compact.

Theorem 14.2.10. Let $X$ be a Hausdorff space and $x \in X$ any point. For every compact set $K \subseteq X \backslash\{x\}$ there exists two disjoint open sets $U$ and $V$ such that $x \in U$ and $K \subseteq V$.

Proof. Since $X$ is Hausdorff and $x \notin K$, for any $k \in K$ there exists disjoint neighbourhoods $U_{x}$ and $V_{k}$ of $x$ and $k$, respectively. Let $\left\{V_{k}\right\}_{k \in K}$ be a collection of such neighbourhoods-which is also an open cover of $K$. Since $K$ is compact, there exists a finite collection of points such that $\left\{V_{k_{1}}, \ldots, V_{k_{n}}\right\}$ is a finite subcover. Defining $V:=\bigcup_{j=1}^{n} V_{k_{j}}$ and $U:=\bigcup_{j=1}^{n} U_{k_{j}}$, we find that $K \subseteq V$ and $x \in U$.

Corollary 14.2.11 (Hausdorff compact subsets). Any compact subset of a Hausdorff space is closed.
Proof. If $X$ is Hausdorff and $C \subseteq X$ is a compact set, let $x \in X \backslash C$ be any point. From Theorem 14.2.10 we know the existence of a neighbourhood $U$ of $x$ that is disjoint from $K$-thus $U \subseteq X \backslash C$ and hence $C$ is closed.

Corollary 14.2.12 (Maps from compact to Hausdorff spaces). Let $X$ be compact and $Y$ Hausdorff. Every continuous map $f: X \rightarrow Y$ is closed and proper. Moreover, the following are consequential properties:
(a) If $f$ is injective, then it is an embedding.
(b) If $f$ is surjective, then it is a quotient map.
(c) If $f$ is bijective, then it is an isomorphism.

Proof. We first prove that $f$ is closed. Let $C \subseteq X$ be any closed set of $X$, which is therefore compact. From Proposition 14.2 .2 we find that $f(C)$ is compact and from Corollary 14.2.11 we have that $f(C)$ is closed.

Now we prove that $f$ is proper. Let $K \subseteq Y$ be any compact set and consider the preimage $f^{-1}(K)$. Since $Y$ is Hausdorff, as pointed before, $K$ is closed. Now since $f$ is continuous, the preimage of closed sets is closed-hence $f^{-1}(K)$ is closed. From the fact that $X$ is compact, we conclude that $f^{-1}(K)$ is compact. For the last three consequences, they come from Proposition 13.5.23.

Lemma 14.2.13 (Tube lemma). Let $X$ be any space and $Y$ be compact. For every $x \in X$ and open set $U \subseteq X \times Y$ containing $\{x\} \times Y$, there exists a neighbourhood $V \subseteq X$ of $x$ such that $V \times Y \subseteq U$.

Proof. Since the product of open sets form a basis for the product topology, for every $y \in Y$ there exists a neighbourhood $V \times W \subseteq U$ of $(x, y)$. Since $\{x\} \times Y \simeq Y$ and $Y$ is compact, then $\{x\} \times Y$ is compact and therefore must exist a finite collection $\left\{V_{j} \times W_{j}\right\}_{j=1}^{n}$ of open sets of $X \times Y$ covering $\{x\} \times Y$. Then if $V:=\bigcap_{j=1}^{n} V_{j}$ we find that $V \times Y \subseteq U$.

Theorem 14.2.14 (Closed projection). Given topological spaces $X$ and $Y$, the space $X$ is compact if and only if the canonical projection $\pi: X \times Y \rightarrow Y$ is closed.

Proof. Suppose $X$ is compact, then if $C \subseteq X \times Y$ is any closed set, let $y \in Y \backslash \pi(C)$ be any point-we shall show that there exists a neighbourhood of $y$ outside of $\pi(C)$. Consider the open set $U:=X \times(Y \backslash \pi(C))$, which certainly contains $X \times\{y\}$. From Lemma 14.2.13, we find a neighbourhood $V \subseteq Y$ of $y$ such that $X \times V \subseteq U$, that is, $V \subseteq Y \backslash \pi(C)$, which settles that $\pi(C) \subseteq Y$ is closed.

Suppose now that $X$ is a space such that $\pi: X \times Y \rightarrow Y$ is closed for any space $Y$. Let $C$ be any collection of closed subsets of $X$ satisfying the finite intersection property-we'll show that $\bigcap_{C \in C} C$ is non-empty. Define $Y$ to be the space to consisting of the underlying set $X \cup\{*\}$ for some point $*$ and the topology given by $2^{X}$ and $\{C \cup\{*\}: C \in C\}$. Let $K \subseteq X \times Y$ be the closure of the diagonal of $X$-that is, $K:=\mathrm{Cl} \Delta_{X}$. From hypothesis, $\pi(K) \subseteq Y$ is closed and from construction $X \subseteq \pi(K)$.

We now show that $* \in \pi(K)$. Suppose on the contrary that $* \notin \pi(K)$. Since $\pi(K)$ is closed, we can find a neighbourhood $V \subseteq Y \backslash \pi(K)$ of $*$ and therefore $V \cap X=\varnothing$. From the construction of the topology of $Y$, one could only hope to write $V$ as the intersection of finitely many sets of the form $C \cup\{*\}$ for $C \in C$-on the other hand, since $C$ satisfies the finite intersection property, for any finite collection of sets of $C$ one can find $x \in X$ such that $x \in C_{1} \cap \cdots \cap C_{n}$, therefore $\left(C_{1} \cup\{*\}\right) \cap \cdots \cap\left(C_{n} \cup\{*\}\right)$ still contains a point of $X$. This shows that it is impossible to build $V$ out of such sets-hence we obtain a contradiction and thus $* \in \pi(K)$ and there exists $\left(x_{0}, *\right) \in K$ for some $x_{0} \in X$.

For every $C \in C$, any neighbourhood $U \times(C \cup\{*\})$ of ( $x_{0}, *$ ) has a non-empty intersection with the diagonal $\Delta_{X}$-thus $C \cap U$ is non-empty. Moreover, $x_{0} \in C$ for all $C \in C$, otherwise, since $C$ is closed, one can find a neighbourhood $U \subseteq X \backslash C$ of $x_{0}$, which should not be possible. Therefore $x_{0} \in \bigcap_{C \in C} C$ and by Proposition 14.2 .7 we conclude that $X$ is compact.

## Tychonoff Theorem

Lemma 14.2.15. Let $J$ be a set, and $\left(X_{j}\right)_{j \in J}$ be a collection of topological spaces. For any point $x \in \prod_{j \in J} X_{j}$ and subset $A \subseteq \prod_{j \in J} X$ of the product space, we have $x \in \mathrm{Cl} A$ if, for every finite $F \subseteq J$, we have $\pi_{F}(x) \in \mathrm{Cl}\left(\pi_{F}(A)\right)$-where $\pi_{F}: \prod_{j \in J} X_{j} \rightarrow \prod_{j \in F} X_{j}$ is the canonical projection map.

Proof. Suppose, on the contrary, that $x \notin \mathrm{Cl} A$-then there exists a neighbourhood $U \subseteq \prod_{j \in J} X_{j}$ of $x$ which is disjoint from $A$. From the definition of the product topology, the collection of preimages $\pi_{j}\left(U_{j}\right)$ for open sets $U_{j} \subseteq X_{j}$ form a sub-basis of the product space. In particular, from the basis properties, this allows for the existence of an open set $V:=U_{j_{1}} \times \cdots \times U_{j_{n}}$ such that, denoting $F:=\left\{j_{1}, \ldots, j_{n}\right\} \subseteq J$, the basis element $\pi_{F}^{-1}(V)$ of the product space is a neighbourhood of $x$ and $\pi_{F}^{-1}(V) \subseteq U$. Therefore $\pi_{F}^{-1}$ and $A$ are disjoint, which is equivalent to

$$
A \subseteq\left(\prod_{j \in J} X_{j}\right) \backslash \pi_{F}^{-1}(V)=\pi_{F}^{-1}\left(\prod_{j \in F} X_{j} \backslash V\right)
$$

that is, $\pi_{F}(A) \subseteq\left(\prod_{j \in F} X_{j}\right) \backslash V$. Since $V$ is a neighbourhood of $\pi_{F}(x)$ which is disjoint from $\pi_{F}(A)$ we conclude, by Proposition 12.1.17, that $\pi_{F}(x) \notin \mathrm{Cl}\left(\pi_{F}(A)\right)$-which proves the lemma.

Theorem 14.2.16 (Tychonoff). The Cartesian product of a set of compact topological spaces, endowed with the product topology, is compact.

Proof. Denote by $\left\{X_{\alpha}\right\}_{\alpha<\kappa}$ a collection of compact spaces indexed by an ordinal $\kappa$. We now show via induction on $\kappa$ that, for any space $Y$, the canonical projection $Y \times \prod_{\alpha<\kappa} X_{\alpha} \rightarrow Y$ is closed.

For the ease of notation we define $X^{\gamma}:=Y \times \prod_{\alpha<\gamma} X_{\alpha}$ for all $\gamma \leqslant \kappa$-moreover for $\lambda \leqslant \gamma$ we denote by $\pi_{\lambda}^{\gamma}: X^{\gamma} \rightarrow X^{\lambda}$ the canonical projection between such spaces. We also define that if $C \subseteq X^{\kappa}$ is a closed set, then $C_{\lambda}:=\mathrm{Cl}\left(\pi_{\lambda}^{\kappa}(C)\right)$-thus our goal is equivalent of showing that $\pi_{0}^{\kappa}(C)=C_{0}$.

Assume as inductive hypothesis that for all $x_{0} \in C_{0}$ there exists $x_{\lambda} \in C_{\lambda}$ for every $\lambda<\kappa$ such that, if $\lambda<\gamma<\kappa$, then

$$
\pi_{\lambda}^{\gamma}\left(x_{\gamma}\right)=x_{\lambda},
$$

and in particular $\pi_{0}^{\lambda}\left(x_{\lambda}\right)=x_{0}$. Again, equivalent to our goal is to show that $\pi_{0}^{\kappa}\left(x_{\kappa}\right)=x_{0}$.
If $\kappa=\lambda+1$ is a successor ordinal, then the projection $\pi_{\lambda}^{\kappa}: X^{\lambda} \times X_{\lambda} \rightarrow X^{\lambda}$ is closed from the fact that $X_{\lambda}$ is compact, by Theorem 14.2.14. In particular, $\pi_{\lambda}^{\kappa}(C) \subseteq X^{\lambda}$ is closed and, by the inductive hypothesis, $\pi_{\lambda}^{\kappa}(C)=C_{\lambda}$-thus there exists $x_{\kappa} \in K$ for which $\pi_{\lambda}^{\kappa}\left(x_{\kappa}\right)=x_{\lambda}$, hence

$$
\pi_{0}^{\kappa}\left(x_{\kappa}\right)=\pi_{0}^{\lambda} \pi_{\lambda}^{\kappa}\left(x_{k}\right)=\pi_{0}^{\lambda}\left(x_{\lambda}\right)=x_{0}
$$

which was out goal.
Now, if $\kappa$ is a limit ordinal, then $X^{\kappa}=\lim _{\lambda<\kappa} X^{\lambda}$ together with transitions maps $\pi_{\lambda}^{\gamma}$. The limit of the tuple $\left(x_{\lambda}\right)_{\lambda<\kappa}$ defines a point $x_{\kappa} \in X^{\kappa}$, we wish to show that $x_{\kappa} \in C$. For every finite set of ordinals $F$ below $\kappa$ there exists $\lambda<\kappa$ above all ordinals of $F$, therefore

$$
\pi_{F}\left(x_{\kappa}\right)=\pi_{F}^{\lambda} \pi_{\lambda}^{\kappa}\left(x_{\kappa}\right)=\pi_{F}^{\lambda}\left(x_{\lambda}\right)
$$

Moreover $\pi_{F}^{\lambda}\left(x_{\lambda}\right) \in \pi_{F}^{\lambda}\left(C_{\lambda}\right)$, where from definition $C_{\lambda}=\operatorname{Cl}\left(\pi_{\lambda}^{\kappa}(C)\right)$. Since $\pi_{F}^{\lambda}$ is a continuous map,

$$
\pi_{F}^{\lambda}\left(C_{\lambda}\right)=\pi_{F}^{\lambda}\left(\mathrm{Cl}\left(\pi_{\lambda}^{\kappa}(C)\right)\right) \subseteq \mathrm{Cl}\left(\pi_{F}^{\lambda} \pi_{\lambda}^{\kappa}(C)\right)=\mathrm{Cl}\left(\pi_{F}(C)\right)
$$

This shows that $\pi_{F}\left(x_{\kappa}\right) \in \operatorname{Cl}\left(\pi_{F}(C)\right)$, which by Lemma 14.2 .15 shows that $\pi(C)$ is closed.

## Applications to Real \& Metric spaces

Lemma 14.2.17 (Closed intervals are compact). Every closed and bounded interval in $\mathbf{R}$ is compact.

Proof. Consider a closed interval $[a, b]$ in $\mathbf{R}$ and let $\mathcal{U}$ be a cover for $[a, b]$. Define $S$ to be the set of all points $x \in[a, b]$ for which the interval $[a, x]$ is covered by finitely many sets of $\mathcal{U}$-since there must exist a set $U \in \mathcal{U}$ for which $a \in U$, then $a \in S$. Since $S$ is non-empty, by the least upper bound property one can define a point $x_{0}:=\sup S$.

Let $U_{0} \in \mathcal{U}$ be a set containing $x_{0}$ and let $\varepsilon>0$ be such that $\left(x_{0}-\varepsilon, x_{0}\right] \subseteq U_{0}$. Since $x_{0}$ is the supremum of $S$, there must also exist $x \in S$ such that $x \in\left(x_{0}-\varepsilon, x_{0}\right]$-that is, the interval $\left[a, x_{0}\right]$ can be covered by finitely many sets of $\mathcal{U}$, say $\left[a, x_{0}\right] \subseteq U_{0} \cup U_{1} \cup \cdots \cup U_{n}$, therefore $x_{0} \in S$. If, for the sake of contradiction, $x_{0}<b$, then by the fact that $U_{0}$ is an open set, there must exist $x \in U_{0}$ with $x>x_{0}$ and yet $x \in[a, b]$-which implies that $[a, x] \subseteq \bigcup_{j=0}^{n} U_{j}$ and therefore $x_{0}$ isn't the supremum of $S$, leading to a contradiction. Thus $x_{0}=b$ and hence $[a, b] \subseteq \bigcup_{j=0}^{n} U_{j}$-which proves the proposition.

Corollary 14.2.18 (Heine-Borel). A subset of $\mathbf{R}^{n}$ is compact if and only if it is both closed and bounded.

Proof. Let $K \subseteq \mathbf{R}^{n}$ be a compact set. If we let $\mathcal{U}$ be the cover of $\mathbf{R}^{n}$ by open balls centred at the origin-with any real radius-in particular $\mathcal{U}$ will be a cover for $K$ and since $K$ is compact, there exists a finite subcover $\mathcal{U}$ covering $K$. From this we conclude that there are only finitely many balls of real radius that cover $K$-hence $K$ is necessarily bounded. From the Hausdorffness of $\mathbf{R}^{n}$, we obtain from Corollary 14.2.11 that $K$ is closed.

On the other hand, let $C \subseteq \mathbf{R}^{n}$ be a closed and bounded set. Since $C$ is a bounded set, its projection into each coordinate must also be bounded, therefore one can obtain a collection of intervals $\left(\left[a_{j}, b_{j}\right]\right)_{j=1}^{n}$ in $\mathbf{R}$ for which

$$
C \subseteq \prod_{j=1}^{n}\left[a_{j}, b_{j}\right] .
$$

From Lemma 14.2.17 we know that each $\left[a_{j}, b_{j}\right]$ is compact—thus by Theorem 14.2.16 we conclude that $\prod_{j=1}^{n}\left[a_{j}, b_{j}\right]$ is compact, but since $C$ is a closed subset of a compact set, by Proposition 14.2 .5 we find that $C$ is compact.

Corollary 14.2.19 (Extreme values on compact sets). Let $X$ be a compact space and $f: X \rightarrow \mathbf{R}$ be a continuous map. Then $f$ is bounded and attains its maximum and minimum values on $X$.

Proof. Since $f(X) \subseteq \mathbf{R}$ is compact, then by Corollary 14.2 .18 we obtain that $f(X)$ is both closed and bounded-thus in particular $f(X)$ contains its supremum and infimum.

We can even extend Corollary 14.2.18 to a more general context, encompassing all metric spaces-this is what the following proposition does.

Proposition 14.2.20. In any metric space $X$, if $A \subseteq X$ is compact then $A$ is both bounded and closed in $X$.

Proof. Since every metric space is Hausdorff, $A$ is closed by Corollary 14.2.11. Moreover, if $x \in A$ is any point, we find that the collection of open balls $\mathcal{B}_{x}:=\left\{B_{x}(n)\right\}_{n \in \mathbf{N}}$ forms an open cover of $A$ and, since $A$ is compact, there exists a finite subcover of $\mathcal{B}_{x}$. Therefore there exists a maximal $m \in \mathbf{N}$ for which $A \subseteq B_{x}(m)$-thus $A$ is bounded.

Remark 14.2.21. Notice that a bounded and closed set in a metric space does not need to be compact, for instance, consider the space $\ell^{2}(\mathbf{N})$ (see Example 26.1.15) and the subset $A:=\left\{f_{n}\right\}_{n \in \mathbf{N}}$ composed of sequences $f_{n} \in \ell^{2}(\mathbf{N})$ such that $f_{n}(j):=\delta_{n j}$-that is, sequences where the only non-zero term is the $n$-th one, which equals 1 . Clearly $A$ is bounded since $\left\|f_{n}\right\|_{2}=1$, moreover, $A$ is closed because it has no limit points. However, since no subset of $A$ contains a limit point of $A$, we find that $A$ is not limit point compact-thus not compact (see Theorem 14.3.8).

## Tychonoff \& The Axiom of Choice

Theorem 14.2.22 (Tychonoff \& Choice). Tychonoff's theorem is equivalent to the axiom of choice.

## Prove

## Examples of Explicit Isomorphisms

Example 14.2.23 (Torus $T^{2}$ ). The 2-torus $T^{2}$ is defined as the quotient $T^{2}:=(I \times I) / \sim$ where $\sim$ is the smallest equivalence relation on $I \times I$ for which $(0, t) \sim(1, t)$ and $(s, 0) \sim(s, 1)$ for all $t, s \in I$. Let's prove that there exists a topological isomorphism

$$
T^{2} \simeq S^{1} \times S^{1}
$$

First, lets consider the interval $I$ and the equivalence relation $0 \sim_{1} 1$ so that $I / \sim_{1}$ is just the interval with its ends glued to a common point. Consider the morphism $f: I \rightarrow S^{1}$ given by $t \mapsto(\cos (2 \pi t), \sin (2 \pi t))$ and notice that $f(1)=f(0)=1$-that is, $f$ is constant on the fibres of the canonical projection $\pi: I \rightarrow I / \sim_{1}$-moreover, $f$ is certainly surjective since cosine and sine are both maps of period $2 \pi$. By Theorem 13.5.3 we find a unique morphism $g: I / \sim_{1} \rightarrow S^{1}$ such that the following diagram commutes


Notice that besides $g$ being surjective, given $[t],[s] \in I / \sim_{1}$ such that $g([t])=g([s])$ then $t-s \in \mathbf{Z}$ since cosine and sine have a period of $2 \pi$-but since $s, t \in I$, this can only be the case if $t=s$, thus $g$ is injective. We therefore conclude that $g$ is a bijective map. Since $I$ is compact, by Corollary 14.2.4 we know that $I / \sim_{1}$ is compact. Moreover, the inclusion map $S^{1} \hookrightarrow \mathbf{R}^{2}$ is continuous and induces the subspace topology on $S^{1}$-since $\mathbf{R}^{2}$ is Hausdorff, then $S^{1}$ is Hausdorff. Therefore $g$ is a bijective map from a compact space $I / \sim_{1}$ to a Hausdorff space $S^{1}$, which by Corollary 14.2.12 is an isomorphism.

Therefore, we find a unique isomorphism $g \times g: T^{2} \xrightarrow{\simeq} S^{1} \times S^{1}$ such that the following diagram commutes

which proves the isomorphism.
Let $\mathbf{R}^{2} \times \mathbf{Z}^{2} \rightarrow \mathbf{R}^{2}$ be a group action of $\mathbf{Z}^{2}$ on $\mathbf{R}^{2}$ given by

$$
((x, y),(n, m)) \mapsto(x+n, y+m)
$$

then the quotient of $\mathbf{R}^{2}$ by such group action, which we'll denote by $\mathbf{R}^{2} / \mathbf{Z}^{2}$ is such that there exists a topological isomorphism

$$
T^{2} \simeq \mathbf{R}^{2} / \mathbf{Z}^{2}
$$

Indeed, one sees right away that $\mathbf{R}^{2} / \mathbf{Z} \simeq(I \times I) / \sim$, where we map $(x, y)+\mathbf{Z}^{2} \mapsto$ $(x-\lfloor x\rfloor, y-\lfloor y\rfloor)$ is the explicit isomorphism. For free we then obtain an isomorphism

$$
S^{1} \times S^{1} \simeq \mathbf{R}^{2} / \mathbf{Z}^{2}
$$

Example 14.2.24 (Projective space $\mathbf{R P}^{2}$ ). Let $\sim_{1}$ and $\sim_{2}$ be the smallest equivalence relations on, respectively, $S^{2}$ and $D^{2}$ such that $p \sim_{1}-p$ for all $p \in S^{2}$, and $v \sim_{2} u$ if $v=\lambda u$ for some $\lambda \in \mathbf{R}$, for all $u, v \in D^{2}$. We'll show that there exists topological isomorphisms

$$
\mathbf{R} \mathbf{P}^{2} \simeq S^{2} / \sim_{1} \simeq D^{2} / \sim_{2}
$$

For the first isomorphism, consider the continuous mapping $f_{1}: \mathbf{R} \mathbf{P}^{2} \rightarrow S^{2} / \sim_{1}$ given by $\mathbf{R} \ell \mapsto\left[\frac{\ell}{\|\ell\|}\right]$ —which collapses the line $\mathbf{R} \ell$ to the class of the unitary vector $\frac{\ell}{\|\ell\|}$ generating the line, which ensures that $f_{1}$ is surjective. Notice that $f_{1}$ is necessarily injective since, if $f_{1}(\mathbf{R} \ell)=f_{1}\left(\mathbf{R} \ell^{\prime}\right)$, then the lines $\mathbf{R} \ell$ and $\mathbf{R} \ell^{\prime}$ intersect both at 0 and at the common point of the sphere-this implies in $\mathbf{R} \ell=\mathbf{R} \ell^{\prime}$. Notice that the inclusion $i_{1}: S^{2} / \sim_{1} \hookrightarrow \boldsymbol{R} \mathbf{P}^{2}$ is continuous and is the inverse of $f_{1}$, sending each unitary vector $v \in S^{2} / \sim_{1}$ of the sphere quotient to the respective line generated by $v$, namely $\mathbf{R} v \in \mathbf{R P}^{2}$.

For the second isomorphism we make an analogous construction, take the map $f_{2}: D^{2} / \sim_{2} \rightarrow S^{2} / \sim_{1}$ given by the identification $[v]_{\sim_{2}} \mapsto\left[\frac{v}{\|v\| \|}\right]_{\sim_{1}}$, which is both continuous and bijective. Moreover, the inclusion $i_{2}: S^{2} / \sim_{1} \hookrightarrow D^{2} / \sim_{2}$ is a continuous map which is inverse to $f_{2}$, thus $f_{2}$ establishes the required isomorphism.

### 14.3 Sequential \& Limit Point Compactness

Definition 14.3.1 (Limit point compactness). A space $X$ is said to be limit point compact if every infinite subset of $X$ has a limit point in $X$.

The following important theorem establishes that every compact space is limit point compact.

Theorem 14.3.2 (Bolzano-Weierstraß). Let $X$ be a compact space. Any infinite subset $S \subseteq X$ has a limit point.

Proof. Suppose, for the sake of contradiction, that there exists $S \subseteq X$, infinite, with no limit points-from this hypothesis, any point $x \in X$ is not a limit point of $S$ and is either in or out of $S$. In the former case $x \in S$ there exists a neighbourhood of $x$, say $U_{x}$, for which $U_{x} \cap S=\{x\}$. In the latter case $x \notin S$, there must exist a neighbourhood $U_{x}$ of $x$ such that $S$ and $U_{x}$ are disjoint. From the law of excluded middle, the collection $\mathcal{U}:=\left\{U_{x}\right\}_{x \in X}$ is an open cover of $X$-on the other hand, there exists no finite subcover of $\mathcal{U}$ since each one must only intersect $S$ at a unique point, but $S$ is infinite, hence the contradiction. The infinite set $S$ must therefore contain at least one limit point.

Remark 14.3.3. One should beware that the converse of Theorem 14.3.2 does not hold at all. A simple counterexample goes as follows: endow $\mathbf{R}$ with the topology given by $\{\varnothing, \mathbf{R}\}$ and the open intervals $\{(x, \infty): x \in \mathbf{R}\}$-in such a space any set has a limit point, although the space itself is not compact.

Definition 14.3.4 (Sequential compactness). A space $X$ is said to be sequentially compact if every sequence of points $\left(x_{j}\right)_{j}$ in $X$, there exists a subsequence $\left(x_{j}^{\prime}\right)_{j} \subseteq\left(x_{j}\right)_{j}$ that converges in $X$.

Lemma 14.3.5 (In first countable Hausdorff spaces). Let $X$ be a first countable Hausdorff space. If $X$ is limit point compact, then $X$ is sequentially compact.

Proof. Suppose $X$ is indeed limit point compact and let $\left(x_{j}\right)_{j \in \mathbf{N}}$ be any sequence of points in $X$ and let $S:=\left\{x_{j}\right\}_{j \in \mathbf{N}}$ be the set of values that the sequence attains. If $S$ is finite, then the sequence contains a constant subsequence-which is therefore convergent in $X$.

On the other hand, if $S$ is infinite, then by the limit point compactness property of $X$ we find that there exists a limit point $x \in X$ of $S$. If it is the case that $x_{j}=x$ for infinitely many $j \in \mathbf{N}$, then the collection of such points form a constant sequencewhich converges to $x$. If the former is not the case, then at most a finite amount of points $x_{j}$ are equal to $x$-therefore, one can discard these points from the sequence and only consider the subsequence $\left(x_{j}^{\prime}\right)_{j} \subseteq\left(x_{j}\right)_{j \in \mathbf{N}}$ such that $x_{j}^{\prime} \neq x$ for all indices $j$. Since $X$ is assumed to be first countable, there must exist a neighbourhood basis $\left(B_{j}\right)_{j \in \mathbf{N}}$ at $x$.

We'll now construct a sequence $\left(x_{j_{i}}\right)_{i \in \mathbf{N}} \subseteq\left(x_{j}^{\prime}\right)_{j}$ such that $x_{j_{i}} \in B_{i}$ and thus $x_{j_{i}} \rightarrow x$. Since $x$ is a limit point, one can choose $x_{j_{0}} \in B_{0}$. For the hypothesis of induction, suppose we have chosen $j_{0}<j_{1}<\cdots<j_{n}$ indices so that $x_{j_{i}} \in B_{i}$. By Proposition 12.4.6 we find that, since $B_{n+1}$ is a neighbourhood of $x$, then $B_{x+1}$ has infinitely many points of $S$-therefore, one can certainly choose $j_{n+1}>j_{n}$ for which $x_{j_{n+1}} \in B_{n+1}$. Thus the sequence $\left(x_{j_{i}}\right)_{i \in \mathbf{N}}$ was successfully constructed so that it converges to $x$.

Lemma 14.3.6. Every sequentially compact metric space is second countable.
Proof. Evoking Proposition 12.5.16 it's sufficient to show that a sequentially compact metric space $M$ is separable. We first show that, for every $\varepsilon>0$, the open cover
composed of $\varepsilon$-balls has a finite subcover. For the sake of contradiction, assume there exists some $\varepsilon_{0}>0$ such that the statement is false.

We now construct a sequence $\left(x_{j}\right)_{j \in \mathbf{N}}$ : choose any $x_{0} \in M$, now since $B_{\varepsilon_{0}}\left(x_{0}\right)$ does not cover $M$, one can choose $x_{1} \in M \backslash B_{\varepsilon_{0}}\left(x_{0}\right)$-since no finite collection of $\varepsilon_{0}$-balls covers $M$, we may proceed indefinitely for each $j \in \mathbf{N}$, always choosing $x_{j} \in M \backslash \bigcup_{i<j} B_{\varepsilon_{0}}\left(x_{i}\right)$.

Since $M$ is sequentially compact, there must exist a convergent subsequence $\left(x_{j_{n}}\right)_{n \in \mathbf{N}}$ of $\left(x_{j}\right)_{j \in \mathbf{N}}$ —with $x_{j_{n}} \rightarrow x$ for some $x \in M$. Since convergent sequences are Cauchy in metric spaces, for large enough $j \in \mathbf{N}$, we have $d\left(x_{j+1}, x_{j}\right)<\varepsilon_{0}$-which implies in $x_{j+1} \in B_{\varepsilon_{0}}\left(x_{j}\right)$, a contradiction by the construction of the sequence. Therefore there must indeed exist a finite collection of $\varepsilon_{0}$-balls that cover $M$.

For the separability of $M$, for each $n \in \mathbf{N}$, define $F_{n}$ to be a finite set of points of $M$ for which the $1 / n$-balls centred at each point of $F_{n}$ covers $M$. Since each $F_{n}$ is finite, the union $F:=\bigcup_{n \in \mathbf{N}} F_{n}$ is countable. Now, if $U \subseteq X$ is a non-empty open subset of $X$ and $p \in U$ is any point and let $B_{\varepsilon}(p) \subseteq U$ be any neighbourhood of $p$. Since $\mathbf{R}$ is archimedean, there must exist $n \in \mathbf{N}$ for which $1 / n<\varepsilon$-hence there exists $q \in F_{n}$ for which $p \in B_{1 / n}(q)$ and $q \in B_{\varepsilon}(p)$. We conclude that $F$ is dense in $M$-thus $M$ is separable.

Proposition 14.3.7 (Metric \& second countable spaces). For second countable spaces and metric spaces, sequential compactness implies compactness.

Proof. Let $X$ be a second countable and sequentially compact space. Let $\mathcal{U}$ be any open cover of $X$. Since every second countable space is Lindelöf (see Proposition 12.5.15), there exists a countable subcover $\mathcal{U}^{\prime}:=\left\{U_{j}\right\}_{j \in \mathbf{N}} \subseteq \mathcal{U}$. For the sake of contradiction, suppose that there is no finite subcover of $\mathcal{U}^{\prime}$-that is, for every $j \in \mathbf{N}$, there exists $x_{j} \in X$ such that $x_{j} \notin \bigcup_{j=0}^{n} U_{j}$. If $\left(x_{j}\right)_{j \in \mathbf{N}}$ is a sequence formed by points with such a property, since $X$ is sequentially compact, there must exist a convergent subsequence $\left(x_{j_{n}}\right)_{n \in \mathbf{N}}$ with $x_{j_{n}} \rightarrow x$ for some point $x \in X$. Let $m \in \mathbf{N}$ be such that $x \in U_{m}$-from the definition of limit point, there must exist only finitely many $x_{j_{n}}$ not contained in $U_{m}$. However, that cannot be the case since, for all $j_{n} \geqslant m$, we must have $x_{j_{n}} \notin \bigcup_{n=0}^{m} U_{n}$ thus in particular $x_{j_{n}} \notin U_{m}$-this implies that $\left(x_{j_{n}}\right)_{n \in \mathbf{N}}$ does not converge to $x$, which is a contradiction.

By Lemma 14.3.6 if $M$ is sequentially compact then it's second countable—therefore, by what was shown above we conclude that $M$ is compact.

Theorem 14.3.8. For metric spaces and second countable Hausdorff spaces, limit point compactness, sequential compactness and compactness are equivalent properties.

Proof. This is the result of Theorem 14.3.2, Lemma 14.3.5 and Proposition 14.3.7. 女
Corollary 14.3.9 (Metric space completeness). Every compact metric space is complete.
Proof. Let $M$ be compact and $\left(x_{j}\right)_{j \in \mathbf{N}}$ be any Cauchy sequence. Since $M$ is compact, then it's also sequentially compact, which implies in the existence of a convergent subsequence $\left(x_{j}^{\prime}\right)_{j} \subseteq\left(x_{j}\right)_{j \in \mathbf{N}}$ for which $x_{j}^{\prime} \rightarrow x$ for some $x \in M$. We'll show that, in fact, $x_{j} \rightarrow x$. Let $B_{\varepsilon}(x)$ be any open ball centred at $x$. Since the sequence is Cauchy,
there exists $N \in \mathbf{N}$ such that $d\left(x_{m}, x_{n}\right)<\varepsilon / 2$ for every $m, n>N$. On the other hand, there exists $M \in \mathbf{N}$ such that $d\left(x_{\ell}^{\prime}, x\right)<\varepsilon / 2$ for every $\ell>M$-therefore, for all $n, \ell>\max (N, M)$ we have $d\left(x_{n}, x\right)<d\left(x_{n}, x_{\ell}^{\prime}\right)+d\left(x_{\ell}^{\prime}, x\right)<\varepsilon$, thus $x_{n} \in B_{\varepsilon}(x)$ for all $n>\max (N, M)$. We conclude that all but finitely many points of $\left(x_{j}\right)_{j \in \mathbf{N}}$ are contained in $B_{\varepsilon}(x)$ for an arbitrary $\varepsilon>0$-therefore $x_{j} \rightarrow x$.

### 14.4 Local Compactness

Definition 14.4.1 (Locally compact spaces). A space $X$ is said to be locally compact if and only if, for every point $x \in X$, there exists a compact set $K \subseteq X$ and a neighbourhood $U \subseteq X$ of $x$ for which $U \subseteq K$.

Definition 14.4.2 (Relatively compact). Given a space $X$, a subset $A \subseteq X$ is said to be relatively compact in X if its closure $\mathrm{Cl} A$ is compact.

Proposition 14.4.3 (Local \& relative compactness in Hausdorff spaces). Let $X$ be a Hausdorff space. The following are equivalent properties:
(a) $X$ is locally compact.
(b) Every point of $X$ has a relatively compact neighbourhood in $X$.
(c) $X$ has a basis of relatively compact open sets.

Proof. From the definition of local and relative compactness, it is clear that (c) implies (b) and that (b) implies (a). We therefore only prove that (a) implies (c). Let $X$ be a locally compact Hausdorff space. Take $x_{0} \in X$ to be any point and let $K \subseteq X$ be a compact set containing a neighbourhood $U \subseteq X$ of $x_{0}$. If we let $\mathcal{B}$ be the collection of all neighbourhoods of $x_{0}$ which are contained in $U$, then $\mathcal{B}$ forms a neighbourhood basis at $x_{0}$. Moreover, since $X$ is Hausdorff, $K$ is closed and every $V \in \mathcal{B}$ is also contained in $K$-therefore $\mathrm{Cl} V \subseteq K$ and since a closed subset of a compact set is compact then $\mathrm{Cl} V$ is compact. We conclude that $V$ is relatively compact in $X$ and thus $\mathcal{B}$ is a basis composed of relatively compact open sets.

Lemma 14.4.4. Let $X$ be a locally compact Hausdorff space. For every point $x \in X$ and neighbourhood $U \subseteq X$ of $x$, there exists a relatively compact neighbourhood $V$ of $x$ for which $\mathrm{Cl} V \subseteq U$.

Proof. Since $X$ is locally compact Hausdorff space, by Proposition 14.4.3 there exists a relatively compact neighbourhood $W$ of $x$. Since $\mathrm{Cl}(W) \backslash U$ is closed in $\mathrm{Cl} W$, it follows that it's compact. Evoking Theorem 14.2.10 one can find disjoint open sets $T$ and $Q$ such that $x \in T$ and $\mathrm{Cl}(W) \backslash U \subseteq Q$. Define a set $V:=T \cap W$, then $\mathrm{Cl} V \subseteq \mathrm{Cl} W$ implies in $\mathrm{Cl} V$ compact-thus $V$ is a relatively compact neighbourhood of $x$, we just need to show that $\mathrm{Cl} V \subseteq U$.

Since $T$ and $Q$ are disjoint, then $T \subseteq X \backslash Q$ and therefore $V \subseteq X \backslash Q$-since $X \backslash Q$ is closed, this implies in $\mathrm{Cl} V \subseteq X \backslash Q$. Therefore, since $\mathrm{Cl} V \subseteq \mathrm{Cl} W$, we find that $\mathrm{Cl} V \subseteq \mathrm{Cl}(W) \backslash Q$. From construction, we have $\mathrm{Cl}(W) \backslash U \subseteq Q$, which implies in $\mathrm{Cl}(W) \backslash Q \subseteq U$-hence in particular $\mathrm{Cl} V \subseteq U$.

Proposition 14.4.5. Any open or closed subset of a locally compact Hausdorff space is itself locally compact Hausdorff space.

Proof. Let $X$ be a locally compact Hausdorff space. If $U \subseteq X$ is an open set, then by Proposition 14.4.3, since $U$ is also Hausdorff, any point $x \in U$ has a relatively compact neighbourhood of $x$ contained in $U$-thus $U$ is locally compact Hausdorff space.

Let $C \subseteq X$ be a closed set and let $x \in C$ be any point. Since $C$ is also Hausdorff, we argue analogously that $x$ has a relatively compact neighbourhood $K$ of $x$ contained in $C$. Since $\mathrm{Cl} K$ is compact, the closed subset $\mathrm{Cl}(K \cap C) \subseteq \mathrm{Cl} K$ is compact. Moreover, since $C$ is closed, $\mathrm{Cl}(K \cap C) \subseteq \mathrm{ClC}=C$, therefore $K \cap C$ is a relatively compact neighbourhood of $x$ in $C$-hence $C$ is a locally compact Hausdorff space.

Lemma 14.4.6 (Product of locally compact spaces). Any finite product of locally compact spaces is locally compact.

Proof. Let $\left\{X_{1}, \ldots, X_{n}\right\}$ be a finite collection of locally compact spaces and consider the product space $X:=\prod_{j=1}^{n} X_{j}$. Let $x \in X$ be any point and consider, for every index $1 \leqslant j \leqslant n$, the projection $\pi_{j}(x) \in X_{j}$. Since $X_{j}$ is locally compact, there exists a compact set $K_{j} \subseteq X_{j}$ containing a neighbourhood $U_{j} \subseteq X_{j}$ of $\pi_{j}(x)$. From this we consider the product sets $K:=\prod_{j=1}^{n} K_{j}$ and $U:=\prod_{j=1}^{n}$. By Theorem 14.2 .16 we know that $K$ is compact and by the product topology the set $U$ is open and also a neighbourhood of $x$. Since $U_{j} \subseteq K_{j}$ for all $1 \leqslant j \leqslant n$, it follows that $U \subseteq K$-which proves that $X$ is locally compact.

Theorem 14.4.7 (Baire category theorem). In every locally compact Hausdorff space or complete metric space, each countable collection of dense open subsets has a dense intersection.

Proof. Let $X$ be either a locally compact Hausdorff space or a complete metric space and $\left\{D_{n}\right\}_{n \in \mathbf{N}}$ be a countable collection of dense open subsets of $X$. Evoking Proposition 12.1.34, it suffices to prove that every non-empty subset $U \subseteq X$ contains a point of $D:=\bigcap_{n \in \mathbf{N}} D_{n}$. We split the cases in two:

1. If $X$ is a locally compact Hausdorff space, we construct a nested sequence of compact sets inductively as follows. From hypothesis, $D_{0}$ is dense, thus in particular $U \cap D_{0}$ contains a point of $X$, therefore evoking Proposition 14.4.3 we can find a non-empty relatively compact open set $C_{0} \subseteq X$ for which $\mathrm{Cl}\left(\mathrm{C}_{0}\right) \subseteq$ $U \cap D_{0}$. We proceed analogously, finding a relatively compact set $C_{1} \subseteq X$ such that $\mathrm{Cl}\left(C_{1}\right) \subseteq C_{0} \cap D_{1} \subseteq U \cap D_{0} \cap D_{1}$. By induction we find a sequence of compact open sets $\left(\mathrm{Cl}\left(C_{n}\right)\right)_{n \in \mathbf{N}}$ such that $\mathrm{Cl}\left(C_{n+1}\right) \subseteq \mathrm{Cl}\left(C_{n}\right)$ and $\mathrm{Cl}\left(C_{n}\right) \subseteq U \cap \bigcap_{j=1}^{n} D_{j}$ for every $n \in \mathbf{N}$. By Lemma 14.2.8, there exists a point $x \in \bigcap_{n \in \mathbf{N}} \mathrm{Cl}\left(C_{n}\right) \subseteq U$ and from construction $x \in \bigcap_{n \in \mathbf{N}} D_{n}$.
2. If $X$ is a complete metric space, we use the inductive argument made above to construct a Cauchy sequence. For every $n \in \mathbf{N}$, we have a non-empty open set $C_{n-1} \cap D_{n}$-hence we may choose a point $x_{n}$ and a neighbourhood $B_{\varepsilon_{n}}\left(x_{n}\right) \subseteq$ $C_{n-1} \cap D_{n}$ for some $\varepsilon_{n}>0$. By choosing for each $n \in \mathbf{N}$ a radius $r_{n}<\min \left(\varepsilon_{n}, 1 / n\right)$,
we construct a sequence of closed balls $\left(\mathrm{Cl}\left(B_{r_{n}}\left(x_{n}\right)\right)\right)_{n \in \mathbf{N}}$ for which $\mathrm{Cl}\left(B_{r_{n}}\left(x_{n}\right)\right) \subseteq$ $U \cap \bigcap_{j=1}^{n} D_{j}$. In the limit $n \rightarrow \infty$ the radius $r_{n}$ of the closed balls goes to zero and $\left(x_{n}\right)_{n \in \mathbf{N}}$ form a Cauchy sequence. Since $X$ is complete, such sequence converges to a point of $U \cap \bigcap_{n \in \mathbf{N}} D_{n}$.

## Manifolds

Definition 14.4.8 (Regular coordinate ball). Let $M$ be an $n$-manifold. We say that a coordinate ball $B \subseteq M$ is regular if there exists a neighbourhood $U$ of $\mathrm{Cl}(B)$ and a topological isomorphism $\phi: U \xrightarrow{\simeq} B_{r^{\prime}}(x)$-for some $r^{\prime}>0$ and $x \in \mathbf{R}^{n}$ —such that $\phi(B)=B_{r}(x)$ and $\phi(\mathrm{Cl} B)=\mathrm{Cl}\left(B_{r}(x)\right)$ for some $0<r<r^{\prime}$.

Corollary 14.4.9. Regular coordinate balls are relatively compact.
Proof. Given a regular coordinate ball $B$, let $U \supseteq B$ be an open set such that there exists an isomorphism $\phi: U \rightarrow B_{r^{\prime}}(x)$. Let $0<r<r^{\prime}$ be such that $\phi(B)=B_{r}(x)$ and $\phi(\mathrm{Cl} B)=\mathrm{Cl}\left(B_{r}(x)\right)$. In particular, the restriction $\left.\phi\right|_{\mathrm{Cl} B}$ is a topological isomorphism, and since $\mathrm{Cl}\left(B_{r}(x)\right)$ is a compact set, so is $\mathrm{Cl} B$.

Lemma 14.4.10. Let $M$ be an $n$-manifold and $B \subseteq M$ be a coordinate ball. If $\phi: B \xrightarrow{\simeq}$ $B_{r^{\prime}}(x)$ is a topological isomorphism—for some $r^{\prime}>0$ and $x \in \mathbf{R}^{n}$-then $\phi^{-1}\left(B_{r}(x)\right) \subseteq M$ is a regular coordinate ball for all $0<r<r^{\prime}$.

Proof. Let $\phi: B \rightarrow B_{r^{\prime}}(x)$ be an isomorphism, and define $U:=\phi^{-1}\left(B_{r}(x)\right)$ for any $0<r<r^{\prime}$. The induced restriction $\phi: U \rightarrow B_{r}(x)$ is clearly an isomorphism.

Consider the restriction $\phi^{-1}: \mathrm{Cl}\left(B_{r}(x)\right) \rightarrow M$. From Corollary 14.2.12 we obtain that $\phi^{-1}$ is a closed map. From Proposition 12.2.12 we find that

$$
\phi^{-1}\left(\operatorname{Cl}\left(B_{r}(x)\right)\right)=\operatorname{Cl}\left(\phi^{-1}\left(B_{r}(x)\right)\right)=\mathrm{Cl}(U)
$$

Therefore $\phi(\mathrm{Cl} U)=\mathrm{Cl}\left(B_{r}(x)\right)$, proving that $U$ is a regular coordinate ball.
Proposition 14.4.11. Every manifold without boundary has a countable basis of regular coordinate balls.

Proof. Let $M$ be any $n$-manifold. Since $M$ is second countable, let $\left\{U_{j}\right\}_{j \in \mathbf{N}}$ be a countable open cover of $M$ consisting of coordinate neighbourhoods. For each $j \in \mathbf{N}$, let $\phi_{j}: U_{j} \xrightarrow{\simeq}$ $V_{j}$ be a topological isomorphism, where $V_{j} \subseteq \mathbf{R}^{n}$ is open. For every $x \in V_{j}$, let $r(x)>0$ be such that $B_{r(x)}(x) \subseteq V_{j}$.

Define $\mathcal{B}$ as the collection of all $\phi_{j}^{-1}\left(B_{r}(x)\right) \subseteq M$ such that $x \in \mathbf{R}^{n}$ is a point consisting of rational coordinates, and $0<r<r(x)$ is also rational. From Lemma 14.4.10 we find that $\phi_{j}^{-1}\left(B_{r}(x)\right)$ is a regular coordinate ball of $M$. Since the rationals are countable, for each $j \in \mathbf{N}$, the collection of regular coordinate balls of the form $\phi_{j}^{-1}\left(B_{r}(x)\right)$ described above is countable. Therefore we conclude that $\mathcal{B}$ is a countable set.

Since each $\phi_{j}$ is an isomorphism, for every $p \in U_{j}$, if $\phi(p)=x$, then $\phi_{j}^{-1}\left(B_{r}(x)\right)$ is a neighbourhood of $p$. Therefore $M=\bigcup_{B \in \mathcal{B}} B$. Moreover, if $B, B^{\prime} \in \mathcal{B}$ are any intersecting sets, let $p \in B \cap B^{\prime}$ be a point. Take $U_{j}$ neighbourhood of $p$ and consider the point $y:=\phi_{j}(p)$. Let $U \subseteq\left(B \cap B^{\prime}\right) \cap U_{j}$ be any neighbourhood of $p$. Let $0<r<r(y)$ be any rational number such that $\phi_{j}(U) \subseteq B_{r}(y)$, then $\phi_{j}^{-1}\left(B_{r}(y)\right) \subseteq U \subseteq B \cap B^{\prime}$ and $\phi_{j}^{-1}\left(B_{r}(y)\right) \in \mathcal{B}$-which proves that $\mathcal{B}$ is a basis of $M$.

Definition 14.4.12 (Regular coordinate half-ball). Let $M$ be an $n$-manifold with boundary. A subset $B \subseteq M$ is a regular coordinate half-ball if there exists an open set $U \supseteq \mathrm{Cl}(B)$ and a topological isomorphism $\phi: U \rightarrow B_{r^{\prime}}(0) \cap \mathbf{H}^{n}$, for some $r^{\prime}>0$, such that $\phi(B) \subseteq B_{r}(0) \cap \mathbf{H}^{n}$ and $\phi(\mathrm{Cl} B) \subseteq \mathrm{Cl}\left(B_{r}(0)\right) \cap \mathbf{H}^{n}$ for some $0<r<r^{\prime}$.

Corollary 14.4.13. Regular coordinate half-balls are relatively compact.
Proof. Follows from the same reasoning of the proof of Corollary 14.4.9.
Proposition 14.4.14. Every manifold with boundary has a countable basis composed of regular coordinate balls and regular coordinate half-balls.

Proof. The proof is analogous to Proposition 14.4.11, one just needs to beware that $V_{j}$ may be either an open subset of $\mathbf{R}^{n}$ or an open subset of $\mathbf{H}^{n}$.

Proposition 14.4.15. Every manifold with or without boundary is locally compact.
Proof. From Corollary 14.4.9 and Proposition 14.4 .11 we find that manifolds without boundary are indeed locally compact.

For manifolds with boundary we analogously use Corollary 14.4.9 and Proposition 14.4.14.

### 14.5 Paracompactness

Definition 14.5.1 (Cover refinement). Let $X$ be a topological space and $\mathcal{U}$ be a cover of $X$. We say that a cover $\mathcal{A}$ of $X$ is a refinement of $\mathcal{U}$ if for each $A \in \mathcal{A}$ there exists $U \in \mathcal{U}$ such that $A \subseteq U$-moreover, $\mathcal{A}$ is said to be an open refinement if each $A \in \mathcal{A}$ is open in $X$.

Corollary 14.5.2. Let $\mathcal{U}$ be an open cover of $X$. If each element of $\mathcal{U}$ intersects only finitely many others, then $\mathcal{U}$ is locally finite.

Proof. Let $p \in X$ be any element. Since $\mathcal{U}$ covers $X$, let $U \in \mathcal{U}$ be a neighbourhood of $p$. From hypothesis, $U$ intersects only finitely many elements of $\mathcal{U}$, therefore $\mathcal{U}$ is locally finite.

Definition 14.5.3 (Paracompactness). A space $X$ is said to be paracompact if for every open cover of $X$ there exists al locally finite open refinement.

Corollary 14.5.4. Every compact space is paracompact.

Proof. Let $X$ be compact and $\mathcal{U}$ be an open cover. If $C \subseteq \mathcal{U}$ is a finite subcover, then in particular $\mathcal{C}$ is a locally finite open refinement of $\mathcal{U}$.

Definition 14.5.5 (Compact exhaustion). Let $X$ be a space. We define a sequence $\left(K_{j}\right)_{j \in \mathbf{N}}$ to be a exhaustion of $X$ by compact sets if $X=\bigcup_{j \in \mathbf{N}} K_{j}$ and, for all $j \in \mathbf{N}$, each set $K_{j}$ is compact in $X$, and the sequence satisfies $K_{j} \subseteq \operatorname{Int} K_{j+1}$.

Lemma 14.5.6. Every second countable, locally compact Hausdorff topological space admits an exhaustion by compact sets.

Proof. Let $X$ be a second countable, locally compact Hausdorff space. Since $X$ is locally compact Hausdorff, there exists a basis consisting of relatively compact open subsets of $X$. Since $X$ is second countable, one can consider a countable cover of $X$ by relatively compact sets $C:=\left\{C_{j}\right\}_{j \in \mathbf{N}}$. We proceed by induction for the construction of each $K_{j}$. Define $K_{0}:=\bar{U}_{0}$, and assume, for the inductive hypothesis, that we've constructed a sequence ( $K_{0}, \ldots, K_{n-1}$ ) of compact sets satisfying $K_{j} \subseteq \operatorname{Int} K_{j+1}$ for $j<n-1$. Since $K_{n-1}$ is compact and $C$ covers $K_{n-1}$, it follows that there exist a finite $k_{n-1} \in \mathbf{N}$ such that $K_{n-1} \subseteq \bigcup_{j=0}^{k_{n-1}} C_{j}$-in particular, choose $k_{n-1}$ so that $k_{n-1}>n$. We thus define $K_{n}:=\bigcup_{j=0}^{k_{n-1}} \mathrm{Cl}\left(C_{j}\right)$ so that $K_{n} \subseteq \operatorname{Int} K_{n-1}$. Moreover, since $k_{n-1}>n$ then $U_{n} \subseteq K_{n}$. Therefore the sequence $\left(K_{j}\right)_{j \in \mathbf{N}}$ is an exhaustion of $X$ by compact sets.

Theorem 14.5.7. Every second countable, locally compact Hausdorff space is paracompact.

Proof. Let $X$ satisfy the above-mentioned hypothesis, and $\mathcal{U}$ be an open cover of $X$. By Lemma 14.5.6, let $\left(K_{j}\right)_{j \in \mathbf{N}}$ be an exhaustion of $X$ by compact sets. Define the following collections:

- For all $j \in \mathbf{N}$, let $A_{j}:=K_{j+1} \backslash \operatorname{Int}\left(K_{j}\right)$ —which is a closed compact set.
- Define $W_{0}:=\operatorname{Int} K_{2}$ and, for each $j>0$, let $W_{j}:=\operatorname{Int}\left(K_{j+2}\right) \backslash K_{j-1}$-forming a collection of open sets.

Notice that $A_{j} \subseteq W_{j}$ for all $j \in \mathbf{N}$.
For each point $p \in A_{j}$, let $U_{p} \in \mathcal{U}$ be a neighbourhood of $p$, and define $V_{p}:=$ $U_{p} \cap W_{j}$-which is again a neighbourhood of $p$. The collection $\left\{V_{p}\right\}_{p \in A_{j}}$ forms an open cover for the compact set $A_{j}$, and from compactness there exists a finite subcover, say $\mathcal{V}_{j}$. The union $\mathcal{V}:=\bigcup_{j \in \mathbf{N}} \mathcal{V}_{j}$ is an open refinement of $\mathcal{U}$.

Consider the open set $W_{j}$ for each $j \in \mathbf{N}$, notice that since $W_{j+3}=\operatorname{Int}\left(K_{j+5}\right) \backslash K_{j+2}$, then $W_{k}$ for $k>j+3$ does not intersect $W_{j}$. On the other hand $W_{j-3}=\operatorname{Int}\left(K_{j-1}\right) \backslash K_{j-4}$ also does not intersect $W_{j}$ and neither do $W_{k}$ for $k<j-3$. Therefore $W_{j}$ intersects $W_{k}$ only for $j-2 \leqslant k \leqslant j+2$-therefore every $V \in \mathcal{V}$ intersects only finitely many members, showing that $\mathcal{V}$ is locally finite.

Corollary 14.5.8. Every manifold with or without boundary is paracompact.

### 14.6 Normal Spaces

Definition 14.6.1 (Normal space). A topological space $X$ is said to be normal if it is Hausdorff, and for every pair of disjoint closed sets $C, F \subseteq X$ there exists a pair of disjoint open sets $U, V \subseteq X$ such that $C \subseteq U$ and $F \subseteq V$.

Proposition 14.6.2. Every compact Hausdorff space is normal.
Proof. Since $X$ is Hausdorff, for each $c \in C$ and $f \in F$ let $C_{f}, F_{f} \subseteq X$ be disjoint neighbourhoods of, respectively, $c$ and $f$. Now consider the collections $\left(C_{f}\right)_{f \in F}$ and $\left(F_{f}\right)_{f \in F}$, where the latter forms an open cover of $F$. Since $F$ is closed, thus compact, let $\mathcal{F}_{c}:=\left(F_{f_{j}}\right)_{j=1}^{n} \subseteq\left(F_{f}\right)_{f \in F}$ be a finite subcover of $F$-and consider the finite intersection $C_{c}:=\bigcap_{j=1}^{n} C_{f_{j}}$, which is a neighbourhood of $c$ and is disjoint from every member of $\mathcal{F}_{c}$.

By the compactness of $C$, let $C:=\left(C_{c_{j}}\right)_{j=1}^{m}$ be a finite subcover of the open cover $\left(C_{c}\right)_{c \in C}$ of the set $C$. If we consider the corresponding family of finite covers $\left(\mathcal{F}_{c_{j}}\right)_{j=1}^{n}$ of the set $F$, we can define $\mathcal{F}:=\bigcap_{j=1}^{m} \mathcal{F}_{c_{j}}$-which is again a finite open cover of $F$, since each $\mathcal{F}_{c_{j}}$ is finite. Now, defining the open sets $U:=\bigcup C$ and $V:=\bigcup \mathcal{F}$, we find that $C \subseteq U$ and $F \subseteq V —$ moreover, from the construction of the open covers, the sets $U$ and $V$ are also disjoint.

Corollary 14.6.3. Every closed subspace of a normal space is normal.
Definition 14.6.4 (Regular space). A topological space $X$ is said to be regular if it is Hausdorff and for every closed set $C \subseteq X$ and point $a \in X \backslash C$, there exists a pair of open sets $U, V \subseteq X$ for which $a \in U$ and $C \subseteq V$.

Lemma 14.6.5. Let $X$ be a Hausdorff space. Then $X$ is normal if and only if, for any closed subset $C \subseteq X$ and open set $U \subseteq X$ containing $C$, there exists a neighbourhood $V$ of $C$ such that $\mathrm{Cl}(V) \subseteq U$.

Proof. If $X$ is normal, we can consider a closed set $C \subseteq X$ and, given an open set $U \subseteq X$ containing $C$, the closed set $F:=X \backslash U$. From definition, there must exist disjoint open sets $V, U^{\prime} \subseteq X$ for which $C \subseteq V$ and $F \subseteq U^{\prime}$. Then in particular $\mathrm{Cl}(V)$ and $F$ are disjoint-otherwise, if $y \in F \cap \mathrm{Cl}(V)$ then $y \in U^{\prime}$, but $U^{\prime}$ and $V$ are disjoint, yielding a contradiction.

For the converse, let $A$ and $B$ be disjoint closed subsets of $X$ and assume the latter hypothesis. Define the open set $U^{\prime}:=X \backslash B$ containing $A$. Then one can choose a neighbourhood $V$ of $A$ such that $\mathrm{Cl}(V) \subseteq U^{\prime}$, therefore $B \subseteq U:=X \backslash \mathrm{Cl}(V)$. Therefore $U$ and $V$ form the desired pair of disjoint open sets for normality.

Theorem 14.6.6. Every paracompact Hausdorff space is normal.
Proof. Let $X$ be a paracompact Hausdorff space, and consider a closed subset $A \subseteq X$ and a point $x \in X \backslash A$-we shall prove that $X$ is regular first. Since $X$ is Hausdorff, for each $p \in A$, let $U_{p}$ and $V_{p}$ be disjoint neighbourhoods of $p$ and $x$, respectively. Therefore $\left(U_{p}\right)_{p \in A}$ forms an open cover of $A$-thus $C^{\prime}:=\left(U_{p}\right)_{p \in P} \cup\{X \backslash A\}$ is an open cover of $X$. Since $X$ is paracompact, let $C$ be a locally finite refinement of $C^{\prime}$, and
consider the subset $\mathcal{U} \subseteq C$ of members $U \subseteq C$ such that $U \subseteq U_{p}$ for some $p \in A$. The collection $\mathcal{U}$ forms a locally finite open cover of $A$. Also, given any $U \in \mathcal{U}$, if $U \subseteq U_{p}$, then $V_{p}$ is a neighbourhood of $x$ that is disjoint from $U$-therefore $\mathrm{Cl}(U)$ cannot contain the point $x$.

Define $L:=\bigcup_{U \in \mathcal{U}} U$ and $W:=X \backslash \mathrm{Cl}(L)$. From the fact that $\mathcal{U}$ is locally finite, we know that $\mathrm{Cl}(L)=\bigcup_{U \in \mathcal{U}} U$. This shows that $x$ is not contained in $L$, while $W$ is a neighbourhood of $x$. Since $L$ and $W$ are disjoint open sets such that $A \subseteq L$ and $x \in W$, this shows that $X$ is regular. For the proof that $X$ is normal, one simply needs to interchange the Hausdorff condition used to build the initial sets by the regularity of $X$.

Corollary 14.6.7. Every topological manifold is normal.
Theorem 14.6.8 (Urysohn). Let $X$ be a normal topological space, and $C, F \subseteq X$ be a pair of disjoint closed subsets. Then there exists a continuous map $f: X \rightarrow[0,1]$ such that $\left.f\right|_{C}=0$ and $\left.f\right|_{F}=1$.

## Part V

## Algebraic Topology

## Chapter 15

## Topological Homotopies \& The Fundamental Group

### 15.1 Homotopy

Definition 15.1.1 (Left homotopy in Top). Let $f, g: X \rightrightarrows Y$ be parallel topological morphisms. We define a left homotopy $\eta: f \Rightarrow g$ between $f$ and $g$ to be a morphism $\eta: X \times I \rightarrow Y$ such that the following diagram commutes

where $i_{0}, i_{1}: X \rightrightarrows X \times I$ are parallel morphisms with $x \stackrel{i_{0}}{\mapsto}(x, 0)$ and $x \stackrel{i_{1}}{\mapsto}(x, 1)$.
If topological spaces are viewed as categories whose objects are its points and morphisms are the identities, then topological morphisms $f, g: X \rightrightarrows Y$ are functors between the categories $X$ and $Y$ and a homotopy $\eta: f \Rightarrow g$ is a natural transformation between $f$ and $g$, where the following diagram commutes in the category $Y$ for any two points $x, x^{\prime} \in X$ :


To put concretely, left homotopy between $f$ and $g$ is a continuous map $\eta$ such that $\eta(-, 0)=f$ and $\eta(-, 1)=g$-which can be visually thought of as a deformation of the morphism $f$ to $g$. If $f$ and $g$ are indeed homotopic, we denote this by $f \sim_{h} g$.

Corollary 15.1.2 (Equivalence relation). Left homotopy induces an equivalence relation $\sim_{h}$ on the collection of topological morphisms. Moreover, homotopy respects composition of morphisms.

Proof. First we prove that $\sim_{h}$ is an equivalence relation:

- (Reflexive) Given a continuous map $f: X \rightarrow Y$, we explicitly define a homotopy $f \Rightarrow f$ by mapping $(x, t) \mapsto x$ for all $(x, t) \in X \times I$.
- (Symmetric) Let $g: X \rightarrow Y$ be another continuous map and suppose that there exists a homotopy $\eta: f \Rightarrow g$. We can define a homotopy $\eta^{\prime}: g \Rightarrow f$ by mapping $(x, t) \mapsto \eta(x, 1-t)$-so that $\eta^{\prime}(-, 0)=\eta(-, 1)=g$ and $\eta^{\prime}(-, 1)=\eta(-, 0)=f$.
- (Transitive) Consider yet another morphism $h: X \rightarrow Y$ and suppose that there exists a homotopy $\sigma: g \Rightarrow h$. In order to construct a homotopy $\delta: f \Rightarrow h$, we define a $X$-parametrized path by concatenating $\eta$ and $\sigma$ appropriately

$$
\delta(x, t):= \begin{cases}\eta(x, 2 t), & t \in[0,1 / 2] \\ \sigma(x, 2 t-1), & t \in[1 / 2,1] .\end{cases}
$$

Indeed, $\delta(-, 0)=\eta(-, 0)=f$ while $\delta(-, 1)=\sigma(-, 1)=h$. It remains to prove that $\delta$ is continuous. Notice that the sets $A:=X \times[0,1 / 2]$ and $B:=X \times[1 / 2,1]$ are both closed in the product topology and cover the whole space $X \times I$. By Proposition 13.2.9, since $\delta$ is continuous on both $A$ and $B$ then $\delta$ is continuous on $A \cup B=X \times I$. We say that $\delta$ is the product of the homotopies $\eta$ and $\sigma$.

To prove that homotopy preserves composition, consider the following commutative diagram in Top

$$
X \xrightarrow{f} Y \xrightarrow[g^{\prime}]{\stackrel{g}{\longrightarrow}} Z \xrightarrow{h} W
$$

If there exists a homotopy $\eta: g \Rightarrow g^{\prime}$, we want to show that $h g f$ is homotopy equivalent to $h g^{\prime} f$. To do that, consider the following commutative diagram


W
From the diagram we see that a natural choice of homotopy $\sigma: h g f \Rightarrow h g^{\prime} f$ is given by $\sigma=h \eta\left(f \times \mathrm{id}_{I}\right)$ thus indeed $h g f \sim_{\mathrm{h}} h g^{\prime} f$.

Given any two topological spaces $X$ and $Y$, we denote the collection of continuous maps $X \rightarrow Y$ up to homotopy equivalence by

$$
[X, Y]:=\operatorname{Mor}_{\mathrm{Top}}(X, Y) / \sim_{\mathrm{h}} .
$$

We call a morphism $[f] \in[X, Y]$ a homotopy class. Moreover, for any three spaces $X, Y, Z \in T o p$, there exists a unique compositional map

$$
[X, Y] \times[Y, Z] \longrightarrow[X, Z]
$$

such that the following diagram commutes


That is, given morphisms $X \xrightarrow{f} Y \xrightarrow{g} Z$, we have a composition of $[f]$ with $[g$ ] uniquely defined as

$$
[g] \circ[f]:=[g f]
$$

Definition 15.1.3 (Homotopy category). We therefore define a category $\mathrm{Ho}(\mathrm{Top})$ composed of topological spaces-which, viewed as an object of Ho (Top), is called a homotopy type-and classes of continuous morphisms between them up to homotopy.

This quotient operation on the category of topological spaces induces a canonically defined projective functor

$$
\mathcal{\kappa}: \mathrm{Top} \longrightarrow \mathrm{Ho}(\mathrm{Top})
$$

Definition 15.1.4 (Homotopy equivalence). Let $X$ and $Y$ be topological spaces. We say that a continuous map $f: X \rightarrow Y$ is a homotopy equivalence of $X$ and $Y$ if there exists a continuous map $g: Y \rightarrow X$ and homotopies $f g \Rightarrow \mathrm{id}_{Y}$ and $g f \Rightarrow \mathrm{id}_{X}$.

If there exists such homotopy equivalence, we write that $X \simeq_{h} Y$-it is to be noted that homotopy equivalences are exactly the isomorphisms in the homotopy category $\mathrm{Ho}(\mathrm{Top})$.

Corollary 15.1.5. Every topological isomorphism is a homotopy equivalence.
Proof. Let $f: X \xrightarrow{\simeq} Y$ be a topological isomorphism. We consider its image under the functor Top $\rightarrow \mathrm{Ho}$ (Top), which we'll name $[f]: X \rightarrow Y$. Then since $f^{-1}: Y \xrightarrow{\simeq} X$ is also a continuous morphism, we can consider its class $\left[f^{-1}\right]$ and notice that $[f]\left[f^{-1}\right]=$ $\left[f f^{-1}\right]=\left[\mathrm{id}_{Y}\right]$ and $\left[f^{-1}\right][f]=\left[f^{-1} f\right]=\left[\mathrm{id}_{X}\right]$. Therefore, there exists two homotopies $f f^{-1} \Rightarrow \operatorname{id}_{Y}$ and $f^{-1} f \Rightarrow \operatorname{id}_{X}$.

Corollary 15.1.6. Homotopy equivalence is an equivalence relation on the class of topological spaces.

Definition 15.1.7 (Relative homotopy). Let $X$ be a space and $A \subseteq X$ be a subspace. If $f, g: X \rightrightarrows Y$ are parallel continuous maps such that $\left.f\right|_{A}=\left.g\right|_{A}$, a homotopy between $f$ and $g$ relative to $A$-if existent-is a homotopy $\eta: f \Rightarrow_{\mathrm{rel} A} g$ such that

$$
\eta(a, t)=f(a)=g(a) \quad \text { for all } a \in A \text { and } t \in I
$$

If $f$ and $g$ are homotopic relative to $A$, we shall also denote this by $f \sim_{\text {rel }} A g$.

## Contractibility

Definition 15.1.8 (Null-homotopy). A morphism of topological spaces $X \rightarrow Y$ is said to be null-homotopic if it is homotopic to a constant map. A homotopy between a morphism and a constant morphism is called a null-homotopy. In particular, a null-homotopy of the identity map id: $X \rightarrow X$ is a contraction of $X$.

Proposition 15.1.9. Let $f: X \rightarrow Y$ and $g: Y \rightarrow Z$ be continuous maps. If either $f$ or $g$ is null-homotopic, then $g f: X \rightarrow \mathrm{Z}$ is null-homotopic.

Proof. Let $f$ be null-homotopic and $\eta: f \Rightarrow c_{p}$ be a homotopy between $f$ and the constant map $c_{p}: X \rightarrow Y$ on $p \in Y$. The composition $g \eta: X \times I \rightarrow Z$ is a continuous map such that $g \eta(-, 0)=g f$ and $g \eta(-, 1)=g c_{p}$, where $g c_{p}=\bar{c}_{g(p)}$ is the constant map on $g(p) \in Z$. Therefore $g \eta$ is a null-homotopy of $g f$.

If $g$ is null-homotopic with a homotopy $\varepsilon: g \Rightarrow c_{z}$ for some constant map $c_{z}: Y \rightarrow Z$ on $z \in Z$. Notice that the map $\varepsilon \circ\left(f \times \mathrm{id}_{I}\right): X \times I \rightarrow Z$ is continuous and for all $x \in X$ we have

$$
\begin{aligned}
& \varepsilon \circ\left(f \times \mathrm{id}_{I}\right)(x, 0)=\varepsilon(f(x), 0)=g f(x), \\
& \varepsilon \circ\left(f \times \mathrm{id}_{I}\right)(x, 1)=\varepsilon(f(x), 1)=c_{z} f(x)=z
\end{aligned}
$$

Therefore establishing a null-homotopy $g f \sim_{h} \bar{c}_{z}$-where $\bar{c}_{z}: X \rightarrow Z$ is constant on $z$.

Proposition 15.1.10. A continuous map $f: S^{n} \rightarrow Y$ is null-homotopic if and only if there exists a continuous map $F: D^{n+1} \rightarrow Y$ such that $\left.F\right|_{S^{n}}=f$.

Proof. First we prove that $\operatorname{Cone}\left(S^{n}\right) \simeq D^{n+1}$. Let $\phi: \operatorname{Cone}\left(S^{n}\right) \rightarrow D^{n+1}$ to be the map given by $[x, t] \mapsto t x$, then $\|t x\|=t\|x\| \leqslant 1$ and $\operatorname{im} \phi \subseteq D^{n+1}$. Moreover, the map is well defined, since $\phi[x, 0]=0$-which is the only uncertain region of the cone. Since the quotient of a compact space is compact, the space Cone $\left(S^{n}\right)$ is compact-moreover, $D^{n+1}$ is Hausdorff. Notice that $\phi$ establishes a bijection between a compact space and a Hausdorff space, thus $\phi$ is an isomorphism.

Suppose $f$ is null-homotopic to a continuous map $c_{p}: S_{n} \rightarrow Y$, for some point $p \in Y$. Let $\eta: c_{p} \Rightarrow f$ be an inverse null-homotopy for $f$. Define $F$ via the isomorphism Cone $\left(S^{n}\right) \simeq D^{n+1}$ as $F(t x):=\eta(x, t)$, where $x \in S^{n}$ and $t \in I$. From this, for every $s \in S^{n}$ we get $F(s)=\eta(s, 1)=f(s)$, and $F(0)=p$.

For the converse, let $F: D^{n+1} \rightarrow Y$ be an extension of $f$. Define a map $\eta: X \times I \rightarrow$ $Y$ given by a collection of maps $\left(\eta_{t}: S^{n} \rightarrow Y\right)_{t \in I}$, where $\eta_{t}(x):=F(t x)$-which is a continuous map-so that the following diagram commutes in Top for all $t \in I$ :

where $\pi_{t}$ is the continuous map $x \mapsto[x, t]$. From this we find that $\eta_{0}(x)=F(0)$ is constant and $\eta_{1}(x)=F(x)=f(x)$ since $x \in S^{n}$.

It just remains to be shown that $\eta$ is continuous. If $U \subseteq Y$ is any open set, let $\left(x_{0}, t_{0}\right) \in \eta^{-1}(U)$ be any pair. Since $F$ is continuous, then there exists $V \subseteq F^{-1}(U)$ neighbourhood of the point $t_{0} x_{0} \in D^{n+1}$. Since $\phi$ is an isomorphism, there exists an open set $P^{\prime} \times T \subseteq \operatorname{Cone}\left(S^{n}\right)$ that is a neighbourhood of the class $\left[x_{0}, t_{0}\right]$, and $P^{\prime} \times T \subseteq \phi^{-1}(V)$. In particular, we know that $\eta_{t_{0}}$ is continuous, therefore there exists a neighbourhood $P \subseteq \eta_{t_{0}}^{-1}\left(P^{\prime} \times T\right)$ of $x_{0}$ in $S^{n}$.

If $(x, t) \in P \times T$ is any point, then from construction we have $F(t x) \in U$-thus $\eta(x, t) \in U$, which shows that $P \times T \subseteq \eta^{-1}(U)$ is a neighbourhood of $\left(x_{0}, t_{0}\right)$, hence $\eta^{-1}(U)$ is open.

Definition 15.1.11 (Contractible space). A space $X$ is said to be contractible if the unique continuous map $X \rightarrow *$ is a homotopy equivalence. From its construction, contractible spaces are the terminal objects of Ho (Top). A null-homotopy of $\mathrm{id}_{X}$ is said to be a contraction of $X$.

Example 15.1.12 (A ball is contractible). Consider the open (or closed) ball $B^{n} \subseteq \mathbf{R}^{n}$. Let $f: B^{n} \rightarrow *$ be the unique continuous map from the ball to the point space, and $\iota: * \rightarrow B^{n}$ be the map $* \mapsto 0$. We define a homotopy $\eta: \iota f \Rightarrow \operatorname{id}_{B^{n}}$ given by $(p, t) \mapsto t p$, and a homotopy $\sigma: f \iota \Rightarrow \operatorname{id}_{*}$ mapping $(*, t) \mapsto *$, since $f \iota=\mathrm{id}_{*}$. Therefore $B^{n}$ is contractible.

Furthermore, since $B^{n} \simeq \mathbf{R}^{n}$ in Top, it follows that $\mathbf{R}^{n} \simeq_{h} B^{n}$ and thus $\mathbf{R}^{n} \simeq_{h} *$-the euclidean space is contractible.

Proposition 15.1.13. A space $X$ is contractible if and only if $\mathrm{id}_{X}$ is null-homotopic.
Proof. Suppose $X$ is contractible, then $f: X \rightarrow *$ is a homotopy equivalence. Let $g: * \rightarrow X$ be a homotopy inverse of $f$, and let $x_{0} \in X$ be the image of $g$-that is $g(*)=x_{0}$ and hence $g f: X \rightarrow X$ is a constant map $g f(x)=x_{0}$. Therefore $\mathrm{id}_{X}$ is homotopic to the constant map $g f$-and the choice of $x_{0}$ was arbitrary.

For the converse, let $x_{0} \in X$ be any point and let $\eta: \mathrm{id}_{X} \Rightarrow c_{x_{0}}$ be a homotopywhere $c_{x_{0}}: X \rightarrow X$ is the constant map $x \mapsto x_{0}$. Consider the unique continuous map $f: X \rightarrow *$ and let $g: * \rightarrow X$ be given by $g(*)=x_{0}$, then $g f=c_{x_{0}} \sim_{h} \operatorname{id}_{X}$ via $\eta$ and $f g=*=\mathrm{id}_{*}$. Therefore $f$ is a homotopy equivalence.

Proposition 15.1.14. If $X$ is contractible, then it is path connected.
Proof. Let $x, y \in X$ be any two points. Since $X$ is contractible, then $\mathrm{id}_{X}$ is nullhomotopic—hence there exists a constant map $c_{p}: X \rightarrow X$ and a homotopy $\eta$ : $\mathrm{id}{ }_{X} \Rightarrow c_{p}$. Notice that by definition $\eta(x, 0)=x$ and $\eta(x, 1)=p$, therefore $\eta(x,-): I \rightarrow X$ is a path from $x$ to $p$-and analogously, $\eta(y,-): I \rightarrow X$ is a path from $y$ to $p$. To construct a path from $x$ to $y$ we can simply concatenate the path $\eta(x,-)$ with the inverse path $\eta^{-1}(y,-)$, which is given by $\eta^{-1}(y, t)=\eta(y, 1-t)$. Concretely, let $\gamma: I \rightarrow X$ be defined by

$$
\gamma(t):= \begin{cases}\eta(x, 2 t), & \text { if } t \in[0,1 / 2] \\ \eta(y, 2-2 t), & \text { if } t \in[1 / 2,1]\end{cases}
$$

which is continuous by Proposition 13.2.9. Then $\gamma$ is a path from $x$ to $y$.
Remark 15.1.15. The converse of Proposition 15.1.14 is not true in general, for instance, the sphere $S^{2}$ is path connected, but not contractible.

Corollary 15.1.16. If $X$ is contractible and $Y$ is path connected, then every pair of parallel morphisms $X \rightrightarrows Y$ is homotopic, in short, $[X, Y]$ has a single point.

Proof. Let $f, g: X \rightrightarrows Y$ be topological morphisms. Since $X$ is contractible, let $\eta$ : $\mathrm{id}_{X} \Rightarrow$ $c_{p}$ be a homotopy from the identity to the constant function $c_{p}: X \rightarrow X$ at the point $p \in$ $X$. Notice that $f \eta(-, 0)=f \operatorname{id}_{X}=f$ and $f \eta(-, 1)=f c_{p}=c_{f(p)}$, where $\bar{c}_{f(p)}: X \rightarrow Y$ is the map constant on $f(p)$. Given any point $y \in Y$, the collection of paths $\operatorname{Path}_{\gamma}(f(p), y)$ is non-empty, thus one can define a collection $\left(\gamma_{x}\right)_{x \in X}$ of paths $\gamma_{x} \in \operatorname{Path}_{\gamma}(f(p), g(x))$. Define a map $\varepsilon: X \times I \rightarrow Y$ given by

$$
\varepsilon(x, t):= \begin{cases}f \eta(x, t), & \text { if } t \in[0,1 / 2]  \tag{b}\\ \gamma_{x}(2 t-1), & \text { if } t \in[1 / 2,1]\end{cases}
$$

then $\varepsilon$ is continuous by Proposition 13.2.9 and defines a homotopy $\varepsilon: f \Rightarrow g$.
Corollary 15.1.17. Let $Y$ be a contractible space. Then for every topological space $X$ any pair of parallel morphisms $X \rightrightarrows Y$ are homotopic, that is, $[X, Y]$ has a single point.

Proof. Let $f, g: X \rightrightarrows Y$ be any two parallel morphisms. Since $Y$ is contractible, there exists $p \in Y$ such that $\eta: \operatorname{id}_{Y} \Rightarrow c_{p}$, where $c_{p}: Y \rightarrow Y$ is the constant map on $p$. Notice that the composition $\eta \circ\left(f \times \mathrm{id}_{I}\right)$ is a continuous map such that, for all $x \in X$ we have

$$
\begin{aligned}
& \eta \circ\left(f \times \mathrm{id}_{I}\right)(x, 0)=\eta(f(x), 0)=\operatorname{id}_{\gamma}(f(x))=f(x), \\
& \eta \circ\left(f \times \mathrm{id}_{I}\right)(x, 1)=\eta(f(x), 1)=c_{p}(f(x))=p .
\end{aligned}
$$

Therefore $\eta \circ\left(f \times \mathrm{id}_{I}\right): X \times I \rightarrow Y$ defines a homotopy $f \sim_{h} c_{p} f$ where $c_{p} f: X \rightarrow Y$ is merely the constant map $x \mapsto p$, which we'll call $\bar{c}_{p}: X \rightarrow Y$. Now, taking the inverse homotopy $(\eta)^{-1}(y, t)=\eta(y, 1-t)$ we obtain a homotopy $c_{p} \sim_{h} \mathrm{id}_{Y}$. We can then consider the continuous map $\eta^{-1} \circ\left(g \times \mathrm{id}_{I}\right): X \times I \rightarrow Y$

$$
\begin{aligned}
& \eta^{-1} \circ\left(g \times \operatorname{id}_{I}\right)(x, 0)=\eta^{-1}(g(x), 0)=c_{p}(g(x))=p \\
& \eta^{-1} \circ\left(g \times \operatorname{id}_{I}\right)(x, 1)=\eta^{-1}(g(x), 1)=\operatorname{id}_{\curlyvee}(g(x))=g(x)
\end{aligned}
$$

therefore $\eta^{-1} \circ\left(g \times \mathrm{id}_{I}\right)$ is a homotopy $c_{p} g=\bar{c}_{p} \sim_{\mathrm{h}} \mathrm{id}_{Y}$. Therefore the concatenation of homotopies $\varepsilon: X \times I \rightarrow Y$ given by

$$
\varepsilon(x, t):= \begin{cases}\eta \circ\left(f \times \operatorname{id}_{I}\right)(x, 2 t), & \text { if } t \in[0,1 / 2] \\ \eta^{-1} \circ\left(g \times \operatorname{id}_{I}\right)(x, 2 t-1), & \text { if } t \in[1 / 2,1]\end{cases}
$$

defines a homotopy $f \sim_{h} \bar{c}_{p} \sim_{h} g$ as wanted.

### 15.2 Mapping Spaces

## Compact-Open Topology

Definition 15.2.1 (Evaluation map). Let $X$ and $Y$ be topological spaces. We define the evaluation map on $X$ and $Y$ to be the set-function:

$$
\text { eval: } \operatorname{Mor}_{\mathrm{Top}}(X, Y) \times X \longrightarrow Y \quad \text { mapping } \quad(f, x) \longmapsto f(x)
$$

Definition 15.2.2 (Admissible topology). A topology on the set of continuous maps $\operatorname{Mor}_{\text {Top }}(X, Y)$ is said to be admissible if the evaluation map eval: $\operatorname{Mor}_{\text {Top }}(X, Y) \times X \rightarrow Y$ is continuous.

Definition 15.2.3 (Compact-open topology). Let $X$ and $Y$ be topological spaces. A pair $(K, U)$-composed of a compact set $K \subseteq X$ and an open set $U \subseteq Y$-defines a a collection of continuous maps

$$
\operatorname{co}(K, U):=\left\{f \in \operatorname{Mor}_{\mathrm{Top}}(X, Y): f(K) \subseteq U\right\}
$$

on the space $\operatorname{Mor}_{\text {Top }}(X, Y)$. We define the compact-open topology on $\operatorname{Mor}_{\text {Top }}(X, Y)$ to be the topology whose subbasis is the collection of sets $\operatorname{co}(K, U)$ for each $K \subseteq X$ compact and $U \subseteq Y$ open. We'll denote the compact-open topology of $\operatorname{Mor}_{T o p}(X, Y)$ by $\operatorname{co}(X, Y)$.

Proposition 15.2.4. The compact-open topology in $\operatorname{Mor}_{T o p}(X, Y)$ is the coarsest between the admissible topologies in the mapping space.

Proof. Suppose $\tau$ is an admissible topology for $\operatorname{Mor}_{\text {Top }}(X, Y)$. We show that for any $\operatorname{co}(K, U) \in \operatorname{co}(X, Y)$, we have $\operatorname{co}(K, U) \in \tau$. Let $k \in K$ be any point and $f \in \operatorname{co}(K, U)$ be any map with $f(K) \subseteq U$. From definition $\operatorname{eval}(f, k)=f(k) \in U$-moreover, since eval is continuous, then $\operatorname{eval}^{-1}(U) \subseteq \operatorname{Mor}_{\text {Top }}(X, Y) \times X$ is open-there must exist a neighbourhood $V_{k} \subseteq\left(\operatorname{Mor}_{\text {Top }}(X, Y), \tau\right)$ of $f$ and a neighbourhood $W_{k} \subseteq K$ of $k$ such that

$$
\operatorname{eval}\left(V_{k} \times W_{k}\right) \subseteq U
$$

The collection $\left(W_{k}\right)_{k \in K}$ is a cover for $K$ —and since $K$ is compact, let $\left\{W_{k_{1}}, \ldots, W_{k_{n}}\right\}$ be a finite subcover. Therefore, for all $1 \leqslant j \leqslant n$ we have eval $\left(V_{k_{j}} \times W_{k_{j}}\right) \subseteq U$. Since open sets are closed under finite intersections, consider the neighbourhood $V:=V_{k_{1}} \cap \cdots \cap V_{k_{n}}$ of $f$. Let $k \in K$ be any point, then there must exist $1 \leqslant j \leqslant n$ such that $k \in W_{k_{j}}$. Therefore if $g \in V$ is any map, we have

$$
g(k)=\operatorname{eval}(g, k) \in \operatorname{eval}\left(V \times W_{k_{j}}\right) \subseteq \operatorname{eval}\left(V_{k_{j}} \times W_{k_{j}}\right) \subseteq U
$$

hence $g(K) \subseteq U$-thus we conclude that $g \in \operatorname{co}(K, U)$ and $V \subseteq \operatorname{co}(K, U)$ is an open subset containing $f$. This shows that $\operatorname{co}(K, U)$ is open in $\left(\operatorname{Mor}_{\text {Top }}(X, Y), \tau\right)$, therefore $\operatorname{co}(K, U) \in \tau$, which proves the proposition.

Proposition 15.2.5. If $X$ is a locally compact Hausdorff space, then the compact-open topology on $\operatorname{Mor}_{\text {Top }}(X, Y)$ is admissible.

Proof. Let $U \subseteq Y$ be any open set, if $\operatorname{eval}^{-1}(U)$ is non-empty, let $(f, x) \in \operatorname{eval}^{-1}(U)$ be any pair. Since $f$ is continuous, then $f^{-1}(U)$ is an open set and there must exist a neighbourhood $W \subseteq X$ of $x$ such that $W \subseteq f^{-1}(U)$-thus $f(W) \subseteq U$. Since $X$ is locally compact Hausdorff, then $W$ is also locally compact Hausdorff and therefore there must exist a relatively compact $V \subseteq X$ neighbourhood of $x$ such that $V \subseteq \mathrm{Cl}(V) \subseteq W$-so that $L:=\operatorname{co}(\mathrm{Cl}(V), U) \times V$ is a neighbourhood of the pair $(f, x)$. In fact, if $(g, y) \in L$ is any pair, then $g(y) \in U$, therefore $(g, y) \in \operatorname{eval}^{-1}(U)$. This implies in $L \subseteq \operatorname{eval}^{-1}(U)$ and proves that eval ${ }^{-1}(U)$ is open in $\operatorname{Mor}_{\text {Top }}(X, Y) \times X$.

Corollary 15.2.6. When $X$ is locally compact Hausdorff, the compact-open topology is the coarsest admissible topology on $\operatorname{Mor}_{\text {Top }}(X, Y)$.

## Exponential Objects

Definition 15.2.7 (Adjoint map). Let $f: X \times Y \rightarrow Z$ be a topological morphism. We define the adjoint map of $f$ to be the continuous map $f^{\wedge}: X \rightarrow \operatorname{Mor}_{\text {Top }}(Y, Z)$ given by $f^{\wedge}(x)(y):=f(x, y)$-this adjoint map is the currying of $f$.

Proposition 15.2.8. Given a continuous map $f: X \times Y \rightarrow Z$, the adjoint map $f^{\wedge}: X \rightarrow$ $\operatorname{Mor}_{\text {Top }}(Y, Z)$ is continuous and, for every $x \in X$, the map $f^{\wedge}(x): Y \rightarrow Z$ is continuous.

Proof. Let $\operatorname{co}(K, U) \subseteq \operatorname{Mor}_{\text {Top }}(Y, Z)$ be an open set in the compact-open topology. Let $f^{\wedge}(x) \in \operatorname{co}(K, U)$ for some $x \in X$-hence $f(\{x\} \times K) \subseteq U$. From tube lemma (see Lemma 14.2.13) there exists $V \subseteq X$ neighbourhood of $x$ such that $V \times K \subseteq f^{-1}(U)$. Therefore $f^{\wedge}(V) \subseteq \operatorname{co}(K, U)$, which shows that $f^{\wedge}(\operatorname{co}(K, U)) \subseteq X$ is an open set. Since this is the case for any element of the subbasis of $\operatorname{Mor}_{\text {Top }}(Y, Z)$, it follows that $f^{\wedge}$ is continuous.

The last assertion is trivial since $f^{\wedge}(x)=f \iota_{x}$, where $\iota_{x}: Y \hookrightarrow X \times Y$ is the injective map $y \mapsto(x, y)$, which is continuous-therefore $f^{\wedge}(x)$ is a continuous map.

The notion of an adjoint map associated with the space $\operatorname{Mor}_{\text {Top }}(X \times Y, Z)$ gives rise to a set-function

$$
\begin{aligned}
\text { curry: } \operatorname{Mor}_{\mathrm{Top}}(X \times Y, Z) & \longrightarrow \operatorname{Mor}_{\mathrm{Top}}\left(X, \operatorname{Mor}_{\mathrm{Top}}(Y, Z)\right) \\
f & f^{\wedge} .
\end{aligned}
$$

Dual to currying is the notion of uncurrying, which is a set-function

$$
\begin{aligned}
\text { uncurry: } \operatorname{Mor}_{\text {Top }}\left(X, \operatorname{Mor}_{\text {Top }}(Y, Z)\right) & \longrightarrow \operatorname{Mor}_{\text {Top }}(X \times Y, Z) \\
\phi & \phi^{\vee}:=\operatorname{eval}_{Y, Z} \circ\left(\phi \times \mathrm{id}_{Y}\right)
\end{aligned}
$$

that is, $\phi^{\vee}(x, y):=\phi(x)(y)$.
Proposition 15.2.9. Let $X$, and $Z$ be any two spaces and $Y$ be a locally compact Hausdorff space. Given a continuous map $g: X \rightarrow \operatorname{Mor}_{\text {Top }}(Y, Z)$, the uncurried map $g^{\vee}: X \times Y \rightarrow Z$ is continuous.

Proof. Let $U \subseteq Z$ be an open set, and let $(x, y) \in\left(g^{\vee}\right)^{-1}(U)$ be any point. Since $g(x) \in \operatorname{Mor}_{\text {Top }}(Y, Z)$ and $y \in(g(x))^{-1}(U)$, there must exist a neighbourhood $W \subseteq Y$ of $y$ such that $W \subseteq(g(x))^{-1}(U)$. Since $Y$ is locally compact Hausdorff, the open set $W$ is also locally compact Hausdorff and therefore there exists a relatively compact neighbourhood $V \subseteq W$ of $y$ with $V \subseteq \mathrm{Cl}(V) \subseteq W$. Thus $g(x)(\mathrm{Cl}(V)) \subseteq U$, therefore $g(x) \in \operatorname{co}(\mathrm{Cl}(V), U)$-that is $g(x)$ is open in $\operatorname{Mor}_{\text {Top }}(Y, Z)$.

Notice that since $g$ is continuous, there exists a neighbourhood $Q \subseteq X$ of $x$ such that $g(Q) \subseteq \operatorname{co}(\mathrm{Cl}(V), U)$. Moreover the neighbourhood $T \times V \subseteq X \times Y$ of $(x, y)$ is contained in $\left(g^{\vee}\right)^{-1}(U)$, since for any $(a, b) \in T \times V$ we have $g^{\vee}(a, b)=g(a)(b) \in U$. Therefore we've shown that $\left(g^{\vee}\right)^{-1}(U) \subseteq X \times Y$ is an open set, which proves that $g^{\vee}$ is continuous.

Using Proposition 15.2.8 and Proposition 15.2.9, we have proven the following corollary.

Corollary 15.2.10. Let $X, Y$ and $Z$ be topological spaces. If $Y$ is a locally compact Hausdorff space, then currying is a set bijection $\operatorname{Mor}_{\text {Top }}(X \times Y, Z) \simeq \operatorname{Mor}_{\text {Top }}\left(X, \operatorname{Mor}_{\text {Top }}(Y, Z)\right)$.

Theorem 15.2.11. Let $X, Y$ and $Z$ be topological spaces. If both $X$ and $Y$ are Hausdorff and $Y$ is also locally compact, then

$$
\operatorname{Mor}_{\text {Top }}(X \times Y, Z) \xrightarrow[\text { curry }]{\simeq} \operatorname{Mor}_{\text {Top }}\left(X, \operatorname{Mor}_{\text {Top }}(Y, Z)\right)
$$

is a topological isomorphism.
Proof. By Corollary 15.2 .10 it is sufficient to prove that curry and uncurry are continuous maps. Under the compact-open topology we know that an open set of the subbasis of $\operatorname{Mor}_{\text {Top }}\left(X, \operatorname{Mor}_{\text {Top }}(Y, Z)\right.$ ) is of the form $\operatorname{co}(K, V)$, for $K \subseteq X$ compact and $V \subseteq \operatorname{Mor}_{\text {Top }}(Y, Z)$ open-that is, there exists an open set $U \subseteq Z$ and compact subset $L \subseteq Y$ such that $V=\operatorname{co}(L, U)$. Therefore $\operatorname{co}(K, V)=\operatorname{co}(K, \operatorname{co}(L, U))$.

Consider any such open set $W:=\operatorname{co}(K, \operatorname{co}(L, U)) \subseteq \operatorname{Mor}_{\text {Top }}\left(X, \operatorname{Mor}_{\text {Top }}(Y, Z)\right)$. Since $K$ and $L$ are compact, then $K \times L \subseteq X \times Y$ is a compact set. If $f \in \operatorname{curry}^{-1}(W)$, then $f(K \times L)=f^{\wedge}(K)(L) \subseteq U$. This implies that $\operatorname{co}(K \times L, U) \subseteq \operatorname{curry}^{-1}(W)$ is a neighbourhood of $f$-therefore curry ${ }^{-1}(W)$ is open in $\operatorname{Mor}_{\text {Top }}(X \times Y, Z)$. From this we conclude that curry is a continuous map.

Consider now any open set $\operatorname{co}(Q, U)$ of the subbasis of $\operatorname{Mor}_{T o p}(X \times Y, Z)$. Define $K:=\pi_{X}(Q)$ and $L:=\pi_{Y}(Q)$, with $Q \subseteq K \times L$-which are compact sets since the image of a compact set under a continuous map is compact. Let $g \in$ uncurry $^{-1}(\operatorname{co}(Q, U))$ be a curried map. If $(x, y) \in Q$, then $g(x)(y)=g^{\vee}(x, y) \in U$-which implies that $\operatorname{co}(K, \operatorname{co}(L, U)) \subseteq$ uncurry $^{-1}(\operatorname{co}(Q, U))$ is a neighbourhood of $g$.

Proposition 15.2.12. Let $Z$ be a locally compact topological space, and $p: X \rightarrow Y$ be a quotient map. Then the product morphism

$$
p \times \mathrm{id}_{Z}: X \times Z \rightarrow Y \times Z
$$

is a quotient map.

Proof. Let $h: Y \times Z \rightarrow W$ be a set-function-where $W$ is a topological space-such that the composition $h \circ\left(p \times \mathrm{id}_{Z}\right)$ is a continuous map. From Proposition 15.2.8 we find that the adjoint $\left(h \circ\left(p \times \mathrm{id}_{Z}\right)\right)^{\wedge}=h^{\wedge} p$ is continuous. Since $p$ is a quotient map, then from the universal property $h^{\wedge}$ is continuous. Now, since $Z$ is locally compact, the uncurried $h$ is continuous by Proposition 15.2.9.

## Linear Homotopy

Definition 15.2.13 (Star-shaped spaces). A subspace $A \subseteq \mathbf{R}^{n}$ is said to be star-shaped with respect to a point $x \in A$ if, for all $y \in A$, the line path $(1-t) x+t y$ is contained in $A$ for all $t \in[0,1]$. A set $C \subseteq \mathbf{R}^{n}$ is said to be convex if and only if it is star-shaped with respect to each of its points.

Proposition 15.2.14. Let $p: X \rightarrow Y$ be a quotient morphism of topological spaces. If $\eta: Y \times I \rightarrow Z$ is a set-function between topological spaces such that the map $\varepsilon: X \times I \rightarrow Z$ given by $(x, t) \mapsto \eta(p(x), t)$ is a homotopy, then $\eta$ is a homotopy.

Proof. Since $p$ is a quotient map and $I$ is compact (hence locally compact), we know from Proposition 15.2.12 that $p \times \mathrm{id}_{I}: X \times I \rightarrow Y \times I$ is a quotient map. Therefore, since $\eta \circ(p \times \mathrm{id})$ is continuous-since $\varepsilon$ is a homotopy-we find that $\eta$ is a continuous map, thus a homotopy.

Definition 15.2.15. Let $X$ be a space and $Y$ be a star-shaped space. We say that a homotopy $\ell: X \times I \rightarrow Y$ is linear if there exists continuous maps $f, g: X \rightrightarrows Y$ such that

$$
\ell(x, t)=(1-t) f(x)+t g(x)
$$

### 15.3 Retractions \& Deformations

## Retractions \& Cofibrations

Definition 15.3.1 (Retract). Let $X$ be a topological space. A subspace $A \subseteq X$ is said to be a retract of $X$ if the inclusion map $\iota: A \hookrightarrow X$ is a split monomorphism in Top-that is, there exists a continuous map $r: X \rightarrow A$ such that $r \iota=\mathrm{id}_{A}$. We call $r$ a retraction of $X$ to $A$.

A subspace $B \subseteq X$ is said to be a weak retract of $X$ if $i: B \hookrightarrow X$ is a split monomorphism in Ho (Top)—that is, there exists a continuous map $r: X \rightarrow A$ such that $r \iota \sim_{h} \operatorname{id}_{A}$.

Definition 15.3.2 (Homotopy extension property). Let $X$ and $Y$ be topological spaces, and $A \subseteq X$ be a subspace. The pair $(X, A)$ is said to have the homotopy extension property with respect to $Y$ if, given continuous maps $g: X \rightarrow Y$ and $\varepsilon: A \times I \rightarrow Y$-such that $g(a)=\varepsilon(a, 0)$ for all $a \in A$-there exists a continuous map $\eta: X \times I \rightarrow Y$ such that $\eta(x, 0)=g(x)$ for all $x \in X$, and $\left.\eta\right|_{A \times I}=\varepsilon$-that is, $\eta$ is an extension of $\varepsilon$. This can all
be summarized by the following commutative diagram in Top:


Corollary 15.3.3 (Extending morphisms). Let $(X, A)$ have the homotopy extension property with respect to $Y$, and consider parallel morphisms $f, f^{\prime}: A \rightrightarrows Y$. If $f$ is homotopic to $f^{\prime}$, and $f$ has an extension to $X$, then $f^{\prime}$ also admits an extension to $X$.

Proof. Let $g: X \rightarrow Y$ be an extension of $f$-that is, $\left.g\right|_{A}=f$ - and $\varepsilon: f \Rightarrow f^{\prime}$ be a homotopy then, from the homotopy extension property of the pair $(X, A)$, there exists a continuous map $\eta: X \times I \rightarrow Y$ which is an extension of $\varepsilon$. From this we conclude that $\eta(-, 1): X \rightarrow Y$ is an extension of $f^{\prime}$.

Definition 15.3.4 (Cofibration). Let $f: Z \rightarrow X$ be a morphism between topological spaces. We say that $f$ is a cofibration if, for any topological space $Y$, and continuous maps $g: X \rightarrow Y$ and $\varepsilon: Z \times I \rightarrow Y$ such that $g(f(z))=\varepsilon(z, 0)$, for all $z \in Z$, there exists a continuous map $\eta: X \times I \rightarrow Y$ such that $\eta(x, 0)=g(x)$ for all $x \in X$, and $\eta(f(z), t)=\varepsilon(z, t)$ for all $z \in Z$ and $t \in I$. This is summarized by the following commutative diagram in Top:


Example 15.3.5. The inclusion $\operatorname{map} A \hookrightarrow X$ is a cofibration if and only if the pair $(X, A)$ has the homotopy extension property every space.

Theorem 15.3.6 (Weakness of retracts). Let $X$ be a topological space and $A \subseteq X$ be a subspace. If the pair $(X, A)$ has the homotopy extension property with respect to $A$, then $A$ is a weak retract of $X$ if and only if $A$ is a retract of $X$.

Proof. We know that if $A$ is a retract of $X$ then in particular $A$ is a weak retract. We now prove the converse. Let $r: X \rightarrow A$ be a weak retraction of $X$ to $A$. Since $r \iota \sim_{h} \mathrm{id}_{A}$, let $\varepsilon: r \iota \Rightarrow \operatorname{id}_{A}$ be a homotopy. Since $(X, A)$ has the homotopy extension property with respect to $A$, it follows that there exists a map $\eta: X \times I \rightarrow A$ extending $\varepsilon$-that is, since $\varepsilon(a, 0)=r(a)$ for all $a \in A$, then $\eta(x, 0)=r(x)$ for all $x \in X$. If we define $r^{\prime}: X \rightarrow A$ to be the map $r^{\prime}:=\eta(-, 1)$, then $r^{\prime}$ is a retraction of $X$ to $A$-since $\left.r^{\prime}\right|_{A}=\operatorname{id}_{A}-$ and $\eta$ establishes a homotopy $r \sim_{h} r^{\prime}$.

## Deformations

Definition 15.3.7 (Deformation). Let $X$ be a space and $A \subseteq X$ be a subspace. We define a deformation of $A$ in $X$ to be a homotopy

$$
\delta: A \times I \longrightarrow X \quad \text { such that } \quad \delta(-, 0)=\operatorname{id}_{A} .
$$

If $\delta(A \times 1) \subseteq B$, for some subspace $B$ of $X$, then $\delta$ is said to be a deformation of $A$ into $B$-and $A$ is said to be deformable in $X$ into $B$.

In particular, the space $X$ is called deformable if there exists a subspace $A \subseteq X$ such that $X$ is deformable in itself into $A$-hence, contractibility of $X$ is equivalent to $X$ being deformable into a point.

Lemma 15.3.8. A space $X$ is deformable into a subspace $A$ if and only if the inclusion $\operatorname{map} A \hookrightarrow X$ admits a right homotopy inverse.

Proof. Suppose that $t: A \hookrightarrow X$ has a right homotopy inverse $f: X \rightarrow A$. Let $\eta$ : $\mathrm{id}_{X} \Rightarrow$ $\iota f$ be a homotopy, then $\eta(-, 0)=\operatorname{id}_{X}$ and $\eta(X, 1)=\iota f(X) \subseteq A$-therefore $\eta$ is a deformation of $X$ into $A$.

For the converse, suppose $X$ is deformable into $A$, and let $\delta: X \times I \rightarrow X$ be such a deformation. Define a map $f: X \rightarrow A$ to be such that $\iota f=\delta(-, 1)$, then $f$ is continuous and $\delta$ establishes a homotopy $\operatorname{id}_{X} \sim_{h} \iota f$ —thus $f$ is a right homotopy inverse of $l$. $\quad$,

Definition 15.3.9 (Deformation retract). Let $X$ be a space, and $A \subseteq X$ a subspace. We define the following concepts:
(a) The subspace $A$ is said to be a weak deformation retract of $X$ if the canonical inclusion $A \hookrightarrow X$ is a homotopy equivalence.
(b) The subspace $A$ is said to be a deformation retract of $X$ if there exists a retraction $r: X \rightarrow A$ of $\iota$-that is, $r \iota=\mathrm{id}_{X}$-such that there exists a homotopy $\iota r \sim_{h} \mathrm{id}_{X}$. A homotopy $\eta: \mathrm{id}_{X} \Rightarrow \iota r$ is called a deformation retraction of $X$ to $A$.
(c) The subspace $A$ is said to be a strong deformation retract of $X$ if there is a retraction $r: X \rightarrow A$ of $\iota$ that is, $r \iota=\mathrm{id}_{X}$-such that there exists a relative homotopy $\iota r \sim_{\operatorname{rel} A}$ $\mathrm{id}_{X}$. A homotopy $\eta: \mathrm{id}_{X} \Rightarrow_{\mathrm{rel} A^{\prime}} \iota r$ is called a strong deformation retraction of $X$ to $A$.

Lemma 15.3.10. Let $A \subseteq X$ be a subspace of a topological space $X$. Then $A$ is a weak deformation retract of $X$ if and only if $A$ is a weak retract of $X$ and $X$ is deformable into $A$.

Proof. From Lemma 15.3.8, there exists a left homotopy inverse of the inclusion $t: A \hookrightarrow$ $X$ if and only if $X$ is deformable into $A$. On the other hand, $A$ is a weak retract of $X$ if and only if $\iota$ admits a right homotopy inverse. Therefore the proposition follows. $\quad \square$

Example 15.3.11. The $n$-th sphere $S^{n}$ is a strong deformation retraction of the punctured euclidean space $\mathbf{R}^{n+1} \backslash 0$. This is realized by the linear homotopy $\delta:\left(\mathbf{R}^{n+1} \backslash 0\right) \times I \rightarrow$ $\mathbf{R}^{n+1} \backslash 0$ given by

$$
\delta(x, t):=(1-t) x+t \frac{x}{\|x\|}
$$

Lemma 15.3.12. Let $X$ be a space, and $A \subseteq X$ be a retract of $X$. If $X$ is deformable into the retract $A$, then $A$ is a deformation retract of $X$.

Proof. Since $A$ is a retract, there exists a left homotopy inverse $r: X \rightarrow A$ of the inclusion $\iota: A \hookrightarrow X$. From hypothesis, $X$ is deformable into $A$, then there exists a right homotopy inverse $f: X \rightarrow A$ of $\iota$ (by Lemma 15.3.8), notice however that

$$
f=\operatorname{id}_{A} f=(r \iota) f=r(\iota f)=r \operatorname{id}_{X}=r .
$$

Therefore we have $\mathrm{id}_{X} \sim_{\mathrm{h}} t r$-thus $A$ is a deformation retract of $X$.
Corollary 15.3.13 (Removing weaknesses via homotopy extensions). If the pair ( $X, A$ ) has the homotopy extension property with respect to the subspace $A$, then $A$ is a weak deformation retract of $X$ if and only if $A$ is a deformation retract of $X$.

Proof. If $A$ is a weak deformation retract of $X$, then by Lemma 15.3.10 we know that $A$ is a weak retract of $X$ and $X$ is deformable into $A$. From Theorem 15.3.6, since $(X, A)$ has the homotopy extension property, then $A$ is a weak retract if and only if it is a retract. Therefore, since $A$ is a retract of $X$ and $X$ is deformable into $A$, by Lemma 15.3.12 we conclude that $A$ is a deformation retract of $X$. The converse is clearly true since a deformation retract is always a weak deformation retract.

Theorem 15.3.14 (Strengthening deformation retracts). Let $X$ be a space and $A \subseteq X$ be a subspace. Define a subspace

$$
Y:=(X \times 0) \cup(A \times I) \cup(X \times 1)
$$

of the cylinder $X \times I$. Then, if $(X \times I, Y)$ has the homotopy extension property with respect to $X$ and $A$ is closed in $X$, then $A$ is a deformation retract of $X$ if and only if $A$ is a strong deformation retract of $X$.

Proof. Since strong deformation retracts are always deformation retracts, we prove the other side of the equivalence. Assume the hypothesis, and let $A$ be a deformation retract of $X$. Let $\delta: \operatorname{id}_{X} \Rightarrow \iota r$ be a deformation retraction of $X$ to $A$-where, as usual, $\iota: A \hookrightarrow X$ is the inclusion and $r: X \rightarrow A$ is the retraction. Define a map $\varepsilon: Y \times I \rightarrow X$ as follows

$$
\varepsilon((x, t), s):= \begin{cases}x, & \text { if } x \in X \text { and } t=0 \\ \delta(x,(1-s) t), & \text { if } x \in A \text { and } t \in I \\ \delta(r(x), 1-s), & \text { if } x \in X \text { and } t=1\end{cases}
$$

The map is indeed well defined since, given any point $a \in A$, one has $\varepsilon((a, 0), s)=a=$ $\delta(a, 0)$-since $\delta$ is a deformation retraction. Moreover, the compatibility of the last two equalities is met because

$$
\varepsilon((a, 1), 1-s)=\delta(a, 1-s)=\delta(r(a), 1-s)
$$

since $r \iota=\mathrm{id}_{A}$. For the continuity of $\varepsilon$, we use the fact that the sets $(X \times 0) \times I,(A \times I) \times I$ and $(X \times 1) \times I$ form a closed cover of $Y \times I$-moreover, since $\varepsilon$ is continuous in each of the sets of such closed cover-by Proposition 13.2.9 we conclude that $\varepsilon$ is continuous.

If $(x, t) \in Y$ is any pair, since $\delta(x, 0)=x$ and

$$
\delta(r(x), 1)=\operatorname{\iota r}(r(x))=r(x)=\delta(x, 1)
$$

then in general $\varepsilon((x, t), 0)=\delta(x, t)$. Using the homotopy extension property of the pair $(X \times I, Y)$ with respect to $X$, there exists an extension

$$
\chi:(X \times I) \times I \rightarrow X
$$

for which $\chi((x, t), 0)=\delta(x, t)$ for all $(x, t) \in X \times I$, and $\left.\chi\right|_{Y \times I}=\varepsilon$. Now define a map $\eta: X \times I \rightarrow X$ given by $\eta(x, t):=\chi((x, t), 1)$, then we obtain

$$
\eta(x, t)= \begin{cases}\varepsilon((x, 0), 1)=x, & \text { if } x \in X \text { and } t=0 \\ \varepsilon((x, t), 1)=\delta(x, 0)=x, & \text { if } x \in A \text { and } t \in I \\ \chi((x, t), 1), & \text { if } x \in X \text { and } t \in I \\ \varepsilon((x, 1), 1)=\delta(r(x), 0)=\iota r(x), & \text { if } x \in X \text { and } t=1\end{cases}
$$

therefore $\eta(-, 0)=\operatorname{id}_{X}$ and $\eta(-, 1)=\iota r$, while $\eta(a, t)=a$ for all $(a, t) \in A \times I$. Therefore $\eta$ establishes a homotopy $\operatorname{id}_{X} \sim_{\operatorname{rel} A} \iota r$, which shows that $A$ is a strong deformation retract of $X$.

## Mapping Cylinder \& Cofibrations

Definition 15.3.15 (Mapping cylinder). Let $f: X \rightarrow Y$ be a topological morphism. We define the mapping cylinder of $f$ to be the pushout

in Top, where $i_{0}: X \rightarrow X \times I$ is the morphism $x \mapsto(x, 0)$-the definition can be "isomorphically" given by replacing $i_{0}$ by $i_{1}: X \rightarrow X \times I$ mapping $x \mapsto(x, 1)$, the resulting pushouts are isomorphic topological spaces.

Put more concretely, the cylinder of $f$ is the topological space

$$
\operatorname{Cyl}(f)=Y \cup_{f}(X \times I)
$$

where one identifies $(x, 0) \sim f(x)$ for all $x \in X$.
Together with the mapping cylinder, we have two distinguished embedding morphisms $\iota_{X}: X \hookrightarrow \operatorname{Cyl}(f)$ with $x \mapsto[x, 0]$, and $\iota_{Y}: Y \hookrightarrow C y l(f)$ mapping $y \mapsto[y]$. Moreover, one has a retraction $r: \operatorname{Cyl}(f) \rightarrow Y$ given by $[x, t] \mapsto f(x)$ for $(x, t) \in X \times I$, while $[y] \mapsto y$ for $y \in Y$.

Theorem 15.3.16. Given a topological morphism $f: X \rightarrow Y$, there exists morphisms $\iota_{X}: X \rightarrow \operatorname{Cyl}(f)$ and $r: \operatorname{Cyl}(f) \rightarrow Y$ such that the following holds:
(a) The diagram

commutes in Top. That is, every continuous map can be factored through an embedding and a retraction of its mapping cylinder.
(b) There exists a relative homotopy

$$
\operatorname{id}_{C y l(f)} \sim_{\operatorname{rel} Y} \iota_{\curlyvee} r .
$$

Thus $t_{Y}$ and $r$ are homotopy equivalences.
(c) The embedding $\iota_{X}$ of $X$ into the cylinder of $f$ is a cofibration.

Proof. Item (a) follows directly from definition: $r \iota_{X}(x)=r[x, 0]=f(x)$ for any $x \in X$. For item (b), consider the map $\eta: \operatorname{Cyl}(f) \times I \rightarrow \operatorname{Cyl}(f)$ given by

$$
\eta(p, s):= \begin{cases}{[x,(1-s) t],} & \text { if } p=[x, t] \text { for }(x, t) \in X \times I, \\ {[y],} & \text { if } p=[y] \text { for } y \in Y .\end{cases}
$$

Then $\eta$ is continuous in both $(X \times I) \times I$ and $Y \times I$, which forms a cover of $\operatorname{Cyl}(f) \times I$ therefore $\eta$ is continuous. Moreover, for any $y \in Y$ we have $\eta([y],-)=[y]$, hence $\eta$ is a homotopy relative to $Y$. Since $\eta([x, t], 0)=[x, t]$ and $\eta([y], 0)=[y]$ then $\eta(-, 0)=\operatorname{id}_{\mathrm{Cyl}(f)}$. On the other hand, $\eta([x, t], 1)=[x, 0]=[f(x)]$ and $\eta([y], 1)=[y]$, that is, $\eta(-, 0)=\iota_{Y} r$. Therefore we can conclude that $\eta$ establishes a relative homotopy $\operatorname{id}_{C y l(f)} \sim_{\text {rel } Y} \iota_{Y} r$.

In order to prove item (c), let $W$ be any topological space and consider continuous maps $g: \operatorname{Cyl}(f) \rightarrow W$ and $\varepsilon: X \times I \rightarrow W$ such that $g \iota_{X}(x)=\varepsilon(x, 0)$ for all $x \in X$. Define a map $\delta: \operatorname{Cyl}(f) \times I \rightarrow W$ given by

$$
\begin{aligned}
\delta([y], s) & :=g[y], \\
\delta([x, t], s) & := \begin{cases}g\left[x, \frac{2 t-s}{2-s}\right], & \text { if } s \leqslant 2 t, \\
\varepsilon\left(x, \frac{s-2 t}{1-t}\right), & \text { if } 2 t \leqslant s,\end{cases}
\end{aligned}
$$

where $y \in Y$ and $x \in X$. Therefore, one has $\delta([x, t], 0)=g[x, t]$ while $\delta([y], 0)=g[y]$. Moreover, $\delta$ is an extension of $\varepsilon$ since $\left.\delta\right|_{X \times I}=\varepsilon$. This shows that $\iota_{X}$ is a cofibration. \&

Lemma 15.3.17. A continuous map $f: X \rightarrow Y$ is a homotopy equivalence if and only if $X$ is a weak deformation retract of the mapping cylinder $\operatorname{Cyl}(f)$.

Proof. From the factorization of Theorem 15.3 .16 we have $f=r \iota_{X}$. Since $\iota_{Y} r \sim_{h} \operatorname{id}_{\mathrm{Cyl}(f)}$, then $\iota_{X} \sim_{h} \iota_{Y} f$. If $f$ is a homotopy equivalence, then let $g: Y \rightarrow X$ be its homotopy inverse. It follows that the map $g r: \operatorname{Cyl}(f) \rightarrow X$ is a homotopy inverse of $\iota_{X}$. Therefore $X$ is a weak deformation retract of $\operatorname{Cyl}(f)$

Conversely, if $X$ is a weak deformation retract of $\operatorname{Cyl}(f)$, let $k: \operatorname{Cyl}(f) \rightarrow X$ be an homotopy inverse of $\iota_{X}$. Then the map $k \iota_{Y}: Y \rightarrow X$ is a homotopy inverse of $f$.

Theorem 15.3.18. Two topological spaces $X$ and $Y$ have the same homotopy type if and only if both can be embedded as a weak deformation retract of a common space $Z$.

Proof. If $f: X \rightarrow Y$ is a homotopy equivalence, then by Lemma 15.3 .17 we find that $X$ is a weak deformation retract of $\operatorname{Cyl}(f)$. Since $Y$ is also a weak deformation retract of $\mathrm{Cyl}(f)$, the statement follows. For the converse, if $i_{X}: X \rightarrow Z$ and $i_{Y}: Y \rightarrow Z$ are embeddings, that are homotopy equivalences, then let $r: Z \rightarrow Y$ be the retract of $i_{Y}$ and define $f:=r i_{X}: X \rightarrow Y$, then $f$ is a homotopy equivalence $X \simeq_{h} Y$.

### 15.4 Fundamental Groupoid \& The Fundamental Group

## Paths

Notation 15.4.1 (Family of paths). Let $X$ be a space and $x, y \in X$ be any two points. We'll denote by $\operatorname{Path}_{X}(x, y)$ the family of paths $\gamma: I \rightarrow X$ with $\gamma(0)=x$ and $\gamma(1)=y$.

Definition 15.4.2 (Operations on paths). Let $X$ be a topological space. We define the following operations on the space of paths of $X$ :

- Given a path $\gamma: I \rightarrow X$ we define the reverse path of $\gamma$ to be a path $\gamma^{-1}: I \rightarrow X$ given by $\gamma^{-1}(t):=\gamma(1-t)$.
- If $p \in \operatorname{Path}_{X}(x, y)$ and $q \in \operatorname{Path}_{X}(y, z)$ are paths in $X$, we define the concatenation of $p$ with $q$ to be a path $q \cdot p: I \rightarrow X$ given by

$$
(q \cdot p)(t):= \begin{cases}p(2 t), & t \in[0,1 / 2] \\ q(2 t-1), & t \in[1 / 2,1]\end{cases}
$$

Yielding a path $q \cdot p \in \operatorname{Path}_{X}(x, z)$.

- Moreover, we define $\operatorname{cons}_{x}: I \rightarrow X$ to be the unique constant path on $x$-that is, $\operatorname{cons}_{x}(t)=x$ for all $t \in I$.

Proposition 15.4.3. If $X$ is a path connected topological space, then any two paths on $X$ are homotopic.

Proof. Let $f, g: I \rightrightarrows X$ be any two paths. Define, for each $t \in I$, a path $\gamma_{t} \in$ Path $_{X}(f(t), g(t))$ which exists since $X$ is path connected. Define a map $\eta: I \times I \rightarrow X$ given by $\eta(t, s):=\gamma_{t}(s)$, then it is clear that $\eta$ is a homotopy between $f$ and $g$.

Definition 15.4.4 (Homotopy relative boundary). Let $X$ be a space, and $\gamma, \gamma^{\prime}: I \rightrightarrows X$ be paths with common endpoints:

$$
\gamma(0)=\gamma^{\prime}(0)=: p_{0} \quad \text { and } \quad \gamma(1)=\gamma^{\prime}(1)=: p_{1}
$$

We define a homotopy relative boundary between $\gamma$ and $\gamma^{\prime}$ to be a homotopy $\eta: \gamma \Rightarrow \gamma^{\prime}$ such that $\eta$ is constant on the endpoints $p_{0}$ and $p_{1}$-that is,

$$
\eta(0,-)=\operatorname{cons}_{p_{0}} \quad \text { and } \quad \eta(1,-)=\operatorname{cons}_{p_{1}}
$$

Proposition 15.4.5. Homotopy relative boundary is an equivalence relation on the family of paths of a topological space.

Proof. Let $X$ be a topological space and $x, y \in X$ be any two points-we'll consider the family $\operatorname{Path}_{X}(x, y)$. The constant homotopy $\gamma \Rightarrow \gamma$ is clearly a homotopy relative boundary, thus the relation is reflexive.

If $\delta \in \operatorname{Path}_{X}(x, y)$ is another path, and $\eta: \gamma \Rightarrow \delta$ is a homotopy relative boundary, then we can construct the a reverse homotopy $\sigma: \delta \Rightarrow \gamma$ as $\sigma(-, t):=\eta(-, 1-t)$. Notice that $\sigma(0, t)=\eta(0,1-t)=\operatorname{cons}_{x}(1-t)$ is constant on $x$, while $\sigma(1, t)=\eta(1,1-t)=$ $\operatorname{cons}_{y}(1-t)$ is constant on $y$. Therefore the relation is symmetric.

Consider yet another path $p \in \operatorname{Path}_{X}(x, y)$ and a homotopy relative boundary $\eta: \delta \Rightarrow p$. We define a map $\varepsilon: I \times I \rightarrow X$ given by

$$
\varepsilon(-, t):= \begin{cases}\sigma(-, 2 t), & t \in[0,1 / 2] \\ \eta(-, 2 t-1), & t \in[1 / 2,1]\end{cases}
$$

This map is continuous by the same argument used in Corollary 15.1.2 and thus establishes a homotopy relative boundary $\varepsilon: \gamma \Rightarrow p$.

## The Fundamental Groupoid

Definition 15.4.6 (Fundamental groupoid). Given a topological space $X$, we define the fundamental groupoid of $X$ to be the category $\Pi_{1}(X)$ whose objects are the points of $X$, and whose morphisms are paths between those points up to homotopy relative boundary. That is, given $x, y \in X$ we have

$$
\operatorname{Mor}_{\Pi_{1}(X)}(x, y)=\operatorname{Path}_{X}(x, y) / \sim_{\text {hrb }}
$$

where $\sim_{\text {hrb }}$ is the equivalence relation on the family of paths $x \rightarrow y$ given by homotopy relative boundary.

Composition of morphisms in $\Pi_{1}(X)$ is naturally defined by the concatenation of paths-in other words, given points $x, y, z \in X$ we have

where, for any paths $\gamma \in \operatorname{Path}_{X}(x, y)$ and $\delta \in \operatorname{Path}_{X}(y, z)$, the concatenation of paths $(\gamma, \delta) \mapsto \delta \cdot \gamma$ induces a concatenation operation on the respective class paths

$$
([\gamma],[\delta]) \longmapsto[\delta] \cdot[\gamma]:=[\delta \cdot \gamma] .
$$

Corollary 15.4.7. $\Pi_{1}(X)$ is a groupoid.
Proof. We show that $\Pi_{1}(X)$ is a category whose morphisms are isomorphisms.

- Given any point $x \in X$ one has an identity map $\left[\operatorname{cons}_{x}\right] \in \operatorname{Mor}_{\Pi_{1}(X)}(x, x)$.
- Concatenation of class paths is unital with respect to constant paths: given any path class $[\gamma] \in \operatorname{Mor}_{\Pi_{1}(X)}(x, y)$, we have

$$
\left[\operatorname{cons}_{y}\right] \cdot[\gamma]=\left[\operatorname{cons}_{y} \cdot \gamma\right]=[\gamma]=\left[\gamma \cdot \operatorname{cons}_{x}\right]=[\gamma] \cdot\left[\operatorname{cons}_{x}\right]
$$

- Concatenation is associative. Let $x, y, z, w \in X$ be any four points and consider class paths $[\alpha]: x \rightarrow y,[\beta]: y \rightarrow z$, and $[\gamma]: z \rightarrow w$.
First we have to show that $\gamma \cdot(\beta \cdot \alpha)$ and $(\gamma \cdot \beta) \cdot \alpha$ are relative boundary homotopic. To that end, define a continuous map $\tau: I \rightarrow I$ by

$$
\tau(t):= \begin{cases}2 t, & t \in[0,1 / 4] \\ t+\frac{1}{4}, & t \in[1 / 4,1 / 2] \\ \frac{t}{2}+\frac{1}{2}, & t \in[1 / 2,1]\end{cases}
$$

Then, one sees right away that

$$
(\gamma \cdot(\beta \cdot \alpha))(\tau(t))=((\gamma \cdot \beta) \cdot \alpha)(t)
$$

for all $t \in I$-hence $\gamma \cdot(\beta \cdot \alpha) \sim_{\text {hrb }}(\gamma \cdot \beta) \cdot \alpha$. From this, we finally obtain the associativity of the path classes,

$$
[\gamma] \cdot([\beta] \cdot[\alpha])=[\gamma \cdot(\beta \cdot \alpha)]=[(\gamma \cdot \beta) \cdot \alpha]=([\gamma] \cdot[\beta]) \cdot[\alpha] .
$$

- Every path class $[\gamma] \in \operatorname{Mor}_{\Pi_{1}(X)}(x, y)$ is an isomorphism, since it has an inverse $\left[\gamma^{-1}\right] \in \operatorname{Mor}_{\Pi_{1}(X)}(y, x)$. Indeed, one has

$$
[\gamma] \cdot\left[\gamma^{-1}\right]=\left[\gamma \cdot \gamma^{-1}\right]=\left[\operatorname{cons}_{y}\right] \quad \text { and } \quad\left[\gamma^{-1}\right] \cdot[\gamma]=\left[\gamma^{-1} \cdot \gamma\right]=\left[\operatorname{cons}_{x}\right] .
$$

Definition 15.4.8 (The category Top*/). A pointed topological space is a pair ( $X, x$ ) consisting of a topological space $X$ together with a base-point $x \in X$. We define a category Top*/ whose objects are pointed topological spaces, and whose morphisms $f:(X, x) \rightarrow(Y, y)$, for any $(X, x),(Y, y) \in$ Top* ${ }^{*}$, are continuous maps $f: X \rightarrow Y$ such that $f(x)=y$. We say that the morphisms of Top ${ }^{* /}$ preserve base-points.

In Top* ${ }^{*}$, we define a pointed homotopy $\eta: f \Rightarrow g$ between parallel morphisms $f, g:(X, x) \rightrightarrows(Y, y)$ to be a homotopy preserving base-points, that is,

$$
\eta(x,-)=\operatorname{cons}_{y} .
$$

Analogously to homotopies in Top, pointed homotopies define an equivalence relation $\sim_{\mathrm{ph}}$ in $\mathrm{Top}^{* /}$. We define the homotopy category of $\mathrm{Top}^{* /}$ to be the category $\mathrm{Ho}\left(\mathrm{Top}^{* /}\right)$ composed of pointed topological spaces and morphisms

$$
\operatorname{Mor}_{\mathrm{Ho}\left(\operatorname{Top}^{*}\right)}((X, x),(Y, y)):=\operatorname{Mor}_{\operatorname{Top}^{*} /}((X, x),(Y, y)) / \sim_{\mathrm{ph}} .
$$

This quotient induces a natural projective functor

$$
\kappa^{* /}: \mathrm{Top}^{* /} \longrightarrow \mathrm{Ho}\left(\mathrm{Top}^{* /}\right) .
$$

## The Fundamental Group

Definition 15.4.9 (Fundamental group). Let $X$ be a topological space and $x \in X$ be any point. We define the fundamental group of $X$ at the base-point $x$ as the family of loops

$$
\pi_{1}(X, x):=\operatorname{Aut}_{\Pi_{1}(X)}(x)
$$

endowed with the operation of concatenation of paths. Therefore, we see that the fundamental group is a functor

$$
\pi_{1}: \text { Top }^{* /} \longrightarrow \text { Grp. }
$$

Proposition 15.4.10. Let $X$ be a path connected space. For every point $x \in X$ the inclusion functor $\pi_{1}(X, x) \rightarrow \Pi_{1} X$ is an equivalence of categories.

Proof. Indeed, notice that $\pi_{1}(X, x)=\operatorname{Aut}_{\Pi_{1} X}(x)$ and since $X$ is a connected space, $\Pi_{1} X$ is a connected groupoid-that is, for every pair of points $x, y \in \Pi_{1} X$, there exists an isomorphism $x \simeq y$ in $\Pi_{1} X$. It follows that $\operatorname{Aut}_{\Pi_{1} X}(x)=\operatorname{sk}\left(\Pi_{1} X\right)$, and therefore the inclusion functor is an equivalence of categories (see Example 1.4.13).

Definition 15.4.11 (Pushforwards in $\left.\pi_{1}\right)$. Let $f:(X, x) \rightarrow(Y, y)$ be a morphism of pointed topological spaces. There exists an induced pushforward

$$
f_{*}: \pi_{1}(X, x) \longrightarrow \pi_{1}(Y, y)
$$

mapping $[\gamma] \mapsto[f \gamma]$, which establishes a group morphism between the fundamental groups of the initial pointed topological spaces.

Proposition 15.4.12. Let $f, g:(X, x) \rightrightarrows(Y, y)$ be morphisms in Top ${ }^{*}$. If there exists a pointed homotopy $\eta: f \Rightarrow g$, then $f_{*}=g_{*}$.

Proof. Let $\gamma$ be a loop at $x$ representing some class of $\pi_{1}(X, x)$. The pointed homotopy $\eta$ naturally induces a homotopy $f \gamma \Rightarrow g \gamma$-thus $[f \gamma]=[g \gamma]$ in $\pi_{1}(Y, y)$-therefore $f_{*}=g_{*}$.

By Proposition 15.4.12 we obtain a factorization of the fundamental group functor through the homotopy category of pointed topological spaces. To put briefly, the following diagram is quasi-commutative


Corollary 15.4.13 (Preserving isomorphisms). If $f:(X, x) \xrightarrow{\simeq}(Y, y)$ is a homotopy equivalence in Top ${ }^{* /}$, then $f_{*}: \pi_{1}(X, x) \xrightarrow{\simeq} \pi_{1}(Y, y)$ is an isomorphism in Grp.

Proof. If $g$ is the homotopy inverse of $f$, then for any $[\gamma] \in \pi_{1}(X, x)$ we have $g_{*} f_{*}[\gamma]=$ $g_{*}[f \gamma]=[g f \gamma]=[\gamma]$, therefore $g_{*} f_{*}=\operatorname{id}_{\pi_{1}(X, x)}$. Analogously we have $f_{*} g_{*}=\operatorname{id}_{\pi_{1}(\gamma, y)}$. Therefore $f_{*}$ is an isomorphism of groups.

## Simply Connected Spaces

Definition 15.4.14 (Simply connected space). A topological space $X$ is said to be simply connected if it is path connected and its fundamental group is trivial, $\pi_{1}(X, x) \simeq 1$ for any base-point $x \in X$.

## Continue on semi-locally simply connected spaces

## Examples of Fundamental Groups

Example 15.4.15 ( $\pi_{1}$ of the euclidean space). Let $x \in \mathbf{R}^{n}$ be any point. Given any loop $\ell: x \rightarrow x$, one can define a homotopy relative boundary $\eta: \ell \Rightarrow \operatorname{cons}_{x}$ given by $\eta(s, t):=(1-t) \ell(s)+t x$. Therefore the fundamental group of the $n$-dimensional euclidean space is trivial,

$$
\pi_{1}(X, x)=* .
$$

Proposition 15.4.16. The fundamental group of a topological manifold is countable.
Proof. Let $\mathcal{B}$ be a countable open cover of $M$ consisting of coordinate balls. Since $M$ has countably many connected components (see Proposition 14.1.29), those connected components are also path components. This implies that any two $B, B^{\prime} \in \mathcal{B}$ are such that the intersection $B \cap B^{\prime}$ are composed of at most countably many path components (since $M$ is locally path connected). Let $\mathcal{X}$ be a collection containing a single point from each $B \cap B^{\prime}$ for any pair $\left(B, B^{\prime}\right) \in \mathcal{B} \times \mathcal{B}$. For every $B \in \mathcal{B}$ and points $x, y \in \mathcal{X}$ with $x, y \in B$, fix a path $\gamma_{(x, y)}^{B} \in \operatorname{Path}_{B}(x, y)$.

Clearly, the fundamental groups of a path connected component of $M$ based at any two points are always isomorphic. Since $\mathcal{X}$ contains at least one point of each path component of $M$, we may choose a base-point $p \in \mathcal{X}$. Let $\Gamma_{p}$ denote the collection of loops based at $p$ that are equal to a finite concatenation of paths of the form $\gamma_{(x, y)}^{B}$, for some $B \in \mathcal{B}$. Since $\mathcal{B}$ is countable, so is $\Gamma_{p}$. We'll settle to prove that every element of $\pi_{1}(M, p)$ can be represented by a loop in $\Gamma_{p}$.

Let $f$ be any loop based at $p$. Consider the collection $\left(f^{-1}(B)\right)_{B \in \mathcal{B}}$, which is an open cover of $I$. Since $I$ is compact, there exists a finite subcover out of such collection. The finite subcover gives rise to a finite collection of numbers

$$
0=: a_{0}<a_{1}<\cdots<a_{k}:=1
$$

for which the closed interval $\left[a_{j-1}, a_{j}\right]$ is contained in some $f^{-1}(B)$.
For each $0<j \leqslant k$, define $f_{j}: I \rightarrow M$ to be the path given by

$$
f_{j}(t):=f\left((1-t) a_{j-1}+t a_{j}\right),
$$

that is, $f_{j}$ is a path $f\left(a_{j-1}\right) \rightarrow f\left(a_{j}\right)$ —and let $B_{j} \in \mathcal{B}$ be such that $\operatorname{im} f_{j} \subseteq B_{j}$. From construction we have that $f\left(a_{j}\right) \in B_{j} \cap B_{j+1}$ for each $0 \leqslant j<k$.

For each $0 \leqslant j<k$ let $g_{j} \in \operatorname{Path}_{B_{j} \cap B_{j+1}}\left(x_{j}, f\left(a_{j}\right)\right)$, where $x_{j} \in \mathcal{X}$ as in the first paragraph—and $x_{0}, x_{k}:=p$, with constant paths $g_{0}, g_{k}:=\operatorname{cons}_{p}$. Notice that

$$
\begin{aligned}
f & \sim_{\mathrm{h}} f_{k} \cdots f_{1} \\
& \sim_{\mathrm{h}} g_{k}^{-1} f_{k}\left(g_{k-1} g_{k-1}^{-1}\right) \cdots\left(g_{2} g_{2}^{-1}\right) f_{2}\left(g_{1} g_{1}^{-1}\right) f_{1} g_{0} \\
& \sim_{\mathrm{h}}\left(g_{k}^{-1} f_{k} g_{k-1}^{-1}\right) \cdots\left(g_{1}^{-1} f_{1} g_{0}\right),
\end{aligned}
$$

where $g_{j}^{-1} f_{j} g_{j-1} \in \operatorname{Path}_{B_{j}}\left(x_{j-1}, x_{j}\right)$ for all $0<j \leqslant k$. Since each $B_{j}$ is simply connected, then $g_{j}^{-1} f_{j} g_{j-1}$ is relative boundary homotopic to chosen path $\gamma_{\left(x_{j-1}, x_{j}\right)}^{B}$. This shows that $f$ is homotopic to a path in $\Gamma_{p}$-therefore $\pi_{1}(M, p)$ is countable.

### 15.5 Mappings $S^{1} \longrightarrow S^{1}$

We shall study the circle via the inclusion $S^{1} \hookrightarrow \mathbf{C}$ mapping $(\cos (2 \pi t), \sin (2 \pi t)) \mapsto e^{2 \pi i t}$ and the identification $q: I \rightarrow S^{1}$ given by $t \mapsto e^{2 \pi i t}$. In general, we shall also make use of the $\operatorname{map} \mathbf{R} \rightarrow S^{1}$ given by $x \mapsto e^{2 \pi i x}$.

Each path $\phi:(I, 0) \rightarrow(\mathbf{R}, 0)$ between pointed spaces, can be lifted to a unique path $\widehat{\phi}:\left(S^{1}, 1\right) \rightarrow\left(S^{1}, 1\right)$ such that the diagram

commutes in Top. Explicitly, we have

$$
\widehat{\phi}_{n}\left(e^{2 \pi \mathrm{i} t}\right)=e^{2 \pi \mathrm{i} \phi_{n}(t)}
$$

Proposition 15.5.1 (Unwinding pointed maps). Let $f:\left(S^{1}, 1\right) \rightarrow\left(S^{1}, 1\right)$ be a pointed continuous map. Then there exists a unique pointed morphism $\phi:(I, 0) \rightarrow(\mathbf{R}, 0)$ such that $f=\widehat{\phi}$.

Proof. Define $h: I \rightarrow S^{1}$ to be the morphism $h:=f q$-which is uniformly continuous since $I$ is a compact space. From the latter properties, one can find a finite partition

$$
0=: t_{0}<t_{1}<\cdots<t_{k}:=1
$$

of $I$, such that $\left|h(t)-h\left(t_{j}\right)\right|<2$ for all $t \in\left[t_{j}, t_{j+1}\right]$, for each $0 \leqslant j<k$. This is done in order to ensure that $h(t) \neq-h\left(t_{j}\right)$-so that the complex logarithm ${ }^{1}$ of the quotient

[^19]$h(t) / h\left(t_{j}\right)$ is well defined. With these conditions being satisfied, we can define a map $\phi: I \rightarrow \mathbf{R}$ given by
$$
\phi(t):=\frac{1}{2 \pi \mathrm{i}}\left(\log \left(\frac{h\left(t_{1}\right)}{h\left(t_{0}\right)}\right)+\cdots+\log \left(\frac{h\left(t_{k}\right)}{h\left(t_{k-1}\right)}\right)+\log \left(\frac{h(t)}{h\left(t_{j}\right)}\right)\right)
$$
for all $t \in\left[t_{j}, t_{j+1}\right]$. Therefore one has
\[

$$
\begin{aligned}
q \phi(t) & =e^{2 \pi i \phi(t)} \\
& =e^{\log \left(\frac{h\left(t_{j}\right)}{h\left(t_{0}\right)}\right)+\cdots+\log \left(\frac{h\left(t_{j}\right)}{h\left(t_{j-1}\right)}\right)+\log \left(\frac{h(t)}{h\left(t_{j}\right)}\right)} \\
& =e^{\log \left(\frac{h\left(t_{1}\right)}{h\left(t_{0}\right)}\right)} \cdots e^{\log \left(\frac{h\left(t_{j}\right)}{h\left(t_{j-1}\right)}\right)} e^{\log \left(\frac{h(t)}{h\left(t_{k}\right)}\right)} \\
& =\frac{h\left(t_{1}\right)}{h\left(t_{0}\right)} \cdots \frac{h\left(t_{j}\right)}{h\left(t_{j-1}\right)} \cdot \frac{h(t)}{h\left(t_{j}\right)} \\
& =\frac{h(t)}{h\left(t_{0}\right)} .
\end{aligned}
$$
\]

Now, from construction we know that $h\left(t_{0}\right)=h(0)=1$, thus $q \phi(t)=h(t)$. Moreover $\phi(0)=0$, which implies that $\phi$ is a continuous pointed map of the form $(I, 0) \rightarrow(\mathbf{R}, 0)$ such that

$$
\widehat{\phi}(t)=f\left(e^{2 \pi i t}\right)
$$

For the uniqueness of $\phi$, suppose that $\psi:(I, 0) \rightarrow(\mathbf{R}, 0)$ is another pointed morphism such that $f=\widehat{\psi}$. Then in particular $\widehat{\phi}=\widehat{\psi}$, which implies in $e^{2 \pi \mathrm{i} \phi(t)}=e^{2 \pi \mathrm{i} \psi(t)}$ for each $t \in I$. This can only be the case if $\phi(t)-\psi(t) \in \mathbf{Z}$. Therefore, since the map $\phi-\psi$ is a continuous map of the form $(I, 0) \rightarrow(\mathbf{Z}, 0)$, it follows that $\phi=\psi$-since $I$ is connected and $\mathbf{Z}$ is discrete.

Theorem 15.5.2 (Unwinding maps). Let $f: S^{1} \rightarrow S^{1}$ be any morphism, then there exists a unique pointed morphism $\phi:(I, 0) \rightarrow(\mathbf{R}, 0)$ such that $f=f(1) \widehat{\phi}$.

Proof. From $f$ we can define a pointed morphism $g:\left(S^{1}, 1\right) \rightarrow\left(S^{1}, 1\right)$ by $g:=f(1)^{-1} f$, so that indeed $g(1)=1$. From Proposition 15.5.1 there exists a unique pointed morphism $\phi:(I, 0) \rightarrow(\mathbf{R}, 0)$ such that $g=\widehat{\phi}$. Therefore, from the definition of $g$ we find that $f=f(1) \widehat{\phi}$.

Lemma 15.5.3. Let $n \in \mathbf{Z}$ and $\phi \in \operatorname{Path}_{\mathbf{R}}(0, n)$. If we consider the linear path $\phi_{n}: I \rightarrow \mathbf{R}$ given by $\phi_{n}(t):=t n$, then there exists a relative homotopy $\widehat{\phi} \sim_{\text {rel }} 1 \widehat{\phi}_{n}$ between the lifted paths in the circle.

Proof. There exists a relative linear homotopy $\ell: \phi \Rightarrow_{\mathrm{rel}} \partial_{I} \phi_{n}$ given by

$$
\ell(s, t):=(1-t) \phi(s)+t \phi_{n}(s) .
$$

Notice that this homotopy can be lifted to $\widehat{\ell}: S^{1} \times I \rightarrow S^{1}$ mapping

$$
\widehat{\ell}\left(e^{2 \pi \mathrm{i} s}, t\right)=e^{2 \pi \mathrm{i}\left((1-t) \phi(s)+t \phi_{n}(s)\right)}=e^{2 \pi \mathrm{i}(1-t) \phi(s)} e^{2 \pi \mathrm{i} t \phi_{n}(s)}
$$

which is a continuous map such that $\widehat{\ell}\left(e^{2 \pi \mathrm{i} s}, 0\right)=e^{2 \pi \mathrm{i} \phi(s)}=\widehat{\phi}\left(e^{2 \pi \mathrm{is}}\right)$ while on the other end $\widehat{\ell}\left(e^{2 \pi \mathrm{is}}, 1\right)=e^{2 \pi \mathrm{i} \phi_{n}(s)}=\widehat{\phi}_{n}\left(e^{2 \pi \mathrm{is}}\right)$. That is, $\widehat{\ell}$ establishes a relative homotopy $\widehat{\phi} \sim_{\text {rel } 1} \widehat{\phi}_{n}$ as we desired.

Proposition 15.5.4. Let $f: S^{1} \rightarrow S^{1}$ be a morphism. Then there exists a unique integer $n \in \mathbf{Z}$ for which $f \sim_{h} \widehat{\phi}_{n}$, where $\phi_{n}: I \rightarrow \mathbf{R}$ is the linear path $\phi_{n}(t):=t n$.
Proof. From Theorem 15.5.2 we find that $f=f(1) \widehat{\phi}$ for a unique pointed map $\phi:(I, 0) \rightarrow$ $(\mathbf{R}, 0)$. Since $\widehat{\phi}:\left(S^{1}, 1\right) \rightarrow\left(S^{1}, 1\right)$ is a pointed continuous map, it follows that

$$
\widehat{\phi}(1)=\widehat{\phi}\left(e^{2 \pi \mathrm{i}}\right)=e^{2 \pi \mathrm{i} \phi(1)}=1
$$

which implies in $\phi(1):=n \in \mathbf{Z}$. From Lemma 15.5 .3 we get $\widehat{\phi} \sim_{\text {rel } 1} \widehat{\phi}_{n}$.
Notice that if $\zeta_{0}:=e^{2 \pi \mathrm{is} s_{0}} \in S^{1}$ is any fixed point, the multiplication map mul: $S^{1} \rightarrow$ $S^{1}$ given by $\operatorname{mul}(\zeta):=\zeta_{0} \zeta$ is a rotation of the point $\zeta$ on the circle-we'll show that mul $\sim_{\mathrm{h}} \operatorname{id}_{\mathcal{S}^{1}}$. Let $\delta: S^{1} \times I \rightarrow S^{1}$ be the map $\delta\left(e^{2 \pi \mathrm{is}}, t\right):=e^{2 \pi \mathrm{i}\left((1-t) s_{0}+s\right)}$, so that $\delta(-, 0)=\mathrm{mul}$ and $\delta(-, 1)=\mathrm{id}_{S^{1}}$.

From the last paragraph we find that

$$
f=f(1) \cdot \widehat{\phi} \sim_{h} \operatorname{id}_{s_{1}} \widehat{\phi}=\widehat{\phi} \sim_{h} \widehat{\phi}_{n}
$$

therefore $f \sim_{h} \widehat{\phi}_{n}$ as wanted.
Definition 15.5.5 (Degree). Given an endomorphism $f: S^{1} \rightarrow S^{1}$, let $\phi: I \rightarrow \mathbf{R}$ be the unique path such that $f=f(1) \widehat{\phi}$-then we define the degree of $f$ to be $\operatorname{deg} f:=\phi(1) \in \mathbf{Z}$.
Lemma 15.5.6. Let $f, g: S^{1} \rightrightarrows S^{1}$ be endomorphisms of the circle. Then $f \sim_{h} g$ if and only if $\operatorname{deg} f=\operatorname{deg} g$.
Proof. $(\Rightarrow)$ Let $\eta: f \Rightarrow g$ be a homotopy. For each $s \in I$ there exists an endomorphism $\eta_{s}:=\eta(-, s): S^{1} \rightarrow S^{1}$ and, from Theorem 15.5.2, there is a unique pointed continuous $\operatorname{map} \phi_{s}:(I, 0) \rightarrow(\mathbf{R}, 0)$ with $\phi_{s}(1) \in \mathbf{Z}$ for which

$$
\begin{equation*}
\eta_{s}=\eta_{s}(1) \widehat{\phi}_{s} . \tag{15.1}
\end{equation*}
$$

We shall construct an explicit equation for $\phi_{s}$ and prove that the mapping $\Phi: I \times I \rightarrow \mathbf{R}$ given by $\Phi(t, s):=\phi_{s}(t)$ is a homotopy.

We proceed as in the proof of Proposition 15.5.1: define a map $\varepsilon: I \times I \rightarrow S^{1}$ by

$$
\varepsilon(t, s):=\eta_{s}\left(e^{2 \pi \mathrm{i} t}\right)=\eta_{s} q(t)
$$

Since $I \times I$ is compact, $\varepsilon$ is uniformly continuous-therefore one can choose a partition

$$
0=: t_{0}<t_{1}<\cdots<t_{k}:=1
$$

of $I$ such that $\left|\varepsilon(t, s)-\varepsilon\left(t_{j}, s\right)\right|<2$ for all $s \in I$ and $t \in\left[t_{j}, t_{j+1}\right]$-this ensures that the complex logarithm of $\varepsilon(t, s) / \varepsilon\left(t_{j}, s\right) \neq-1$ is well defined. For each $s \in I$, we construct a map $\psi_{s}: I \rightarrow S^{1}$ by

$$
\begin{equation*}
\psi_{s}(t):=\frac{1}{2 \pi \mathrm{i}}\left(\log \left(\frac{\varepsilon\left(t_{1}, s\right)}{\varepsilon\left(t_{0}, s\right)}\right)+\cdots+\log \left(\frac{\varepsilon\left(t_{j}, s\right)}{\varepsilon\left(t_{j-1}, s\right)}\right)+\log \left(\frac{\varepsilon(t, s)}{\varepsilon\left(t_{j}, s\right)}\right)\right) \tag{15.2}
\end{equation*}
$$

for all $t \in\left[t_{j}, t_{j+1}\right]$. Then for every $s \in I$ one has

$$
\begin{aligned}
q \psi_{s}(t) & =e^{2 \pi \mathrm{i} \psi_{s}(t)} \\
& \left.=e^{\log \left(\frac{\varepsilon\left(t_{1}, s\right)}{\varepsilon\left(t_{0}, s\right)}\right)}\right)+\cdots+\log \left(\frac{\varepsilon\left(t_{j}, s\right)}{\varepsilon\left(t_{j-1}, s\right)}\right)+\log \left(\frac{\varepsilon(t, s)}{\varepsilon\left(t_{j}, s\right)}\right) \\
& =e^{\log \left(\frac{\varepsilon\left(t_{1, s}\right)}{\varepsilon\left(t_{0}, s\right)}\right)} \cdots e^{\log \left(\frac{\varepsilon\left(t_{j}, s\right)}{\varepsilon\left(t_{j}-1, s\right)}\right)} e^{\log \left(\frac{\varepsilon(t, s)}{\varepsilon\left(t_{j}, s\right)}\right)} \\
& =\frac{\varepsilon\left(t_{1}, s\right)}{\varepsilon\left(t_{0}, s\right)} \cdots \frac{\varepsilon\left(t_{j}, s\right)}{\varepsilon\left(t_{j-1}, s\right)} \cdot \frac{\varepsilon(t, s)}{\varepsilon\left(t_{j}, s\right)} \\
& =\frac{\varepsilon(t, s)}{\varepsilon\left(t_{0}, s\right)} \\
& =\frac{\eta_{s}\left(e^{2 \pi i t}\right)}{\eta_{s}(1)}
\end{aligned}
$$

Therefore $\eta_{s}=\eta_{s}(1) \widehat{\psi}_{s}$, which shows that $\psi_{s}=\phi_{s}$ (from Eq. (15.1)). Looking at Eq. (15.2) we see that the mapping $\Phi$ above-mentioned is continuous, hence a homotopy.

Considering the continuous map $\Phi(1,-): I \rightarrow \mathbf{R}$ given by $s \mapsto \phi_{s}(1)$, we know that $\phi_{s}(1) \in \mathbf{Z}$ from earlier considerations-therefore using the fact that $I$ is connected and $\mathbf{Z}$ is discrete, it must bae the case that the induced map $\Phi(1,-): I \rightarrow \mathbf{Z}$ is constant. In particular, we shall have $\phi_{0}(1)=\phi_{1}(1)$-but from definition we have $f=f(1) \widehat{\phi}_{0}$ and $g=g(1) \widehat{\phi}_{1}$, so that $\operatorname{deg} f=\phi_{0}(1)$ and $\operatorname{deg} g=\phi_{1}(1)$. From these considerations we can finally conclude that $\operatorname{deg} f=\operatorname{deg} g$.
$(\Leftarrow)$ For the converse, suppose that $\operatorname{deg} f=\operatorname{deg} g:=n$. Therefore there exists two unique pointed morphisms $\phi, \psi \in \operatorname{Path}_{\mathbf{R}}(0, n)$ such that $f=f(1) \widehat{\phi}$ and $g=g(1) \widehat{\psi}$. Recall that since $f(1), g(1) \in S^{1}$ are unitary complex numbers, multiplying a circle point by them amounts to a rotation of the initial point through the circle-which was already shown to be homotopic to the identity map of the circle. Therefore we conclude that $\phi \sim_{h} \phi_{n} \psi$, where $\phi_{n} \in \operatorname{Path}_{\mathbf{R}}(0, n)$ is the linear path $\phi_{n}(t):=t n$. Thus

$$
f \sim_{\mathrm{h}} \widehat{\phi} \sim_{\mathrm{h}} \widehat{\phi}_{n} \sim_{\mathrm{h}} \widehat{\psi} \sim_{\mathrm{h}} g
$$

proving that $f$ and $g$ are homotopic.
Lemma 15.5.7. For each $n \in \mathbf{Z}$ the map $e_{n}: S^{1} \rightarrow S^{1}$ given by $\zeta \mapsto \zeta^{n}$ is a continuous map of degree

$$
\operatorname{deg} e_{n}=n
$$

Proof. Notice that from definition we have $e_{n}\left(e^{2 \pi i t}\right)=e^{2 \pi \mathrm{i}(n t)}=q \phi_{n}(t)$, where $\phi_{n} \in$ $\operatorname{Path}_{\mathbf{R}}(0, n)$ is the linear path $\phi_{n}(t):=t n$. Therefore $\operatorname{deg} e_{n}=\phi_{n}(1)=n$.
Theorem 15.5.8 (deg is a ring isomorphism). The map deg: $\left[S^{1}, S^{1}\right] \rightarrow \mathbf{Z}$ given by $[f] \mapsto \operatorname{deg} f$ is a ring isomorphism. In other words, the degree map is an isomorphism of the first cohomology group of $S^{1}$ with $\mathbf{Z}$ :

$$
H^{1}\left(S^{1}\right) \simeq \mathbf{Z} .
$$

Proof. From Lemma 15.5 .6 we know that deg is a bijective map, we need to show that it's also a ring morphism. Consider any two endomorphism classes $[f],[g] \in\left[S^{1}, S^{1}\right]$ with degrees $\operatorname{deg}[f]:=n$ and $\operatorname{deg}[g]:=m$, then:

- (Additive structure). From Lemma 15.5 .7 we know that $f \sim_{h} e_{n}$ and $g \sim_{h} e_{m}$, moreover given any $\zeta \in S^{1}$ one has

$$
\left(e_{n} \cdot e_{m}\right)(\zeta)=e_{n}(\zeta) e_{m}(\zeta)=\zeta^{n+m}
$$

Therefore $e_{n} \cdot e_{m}=e_{n+m}$ and, consequently

$$
\operatorname{deg}([f] \cdot[g])=\operatorname{deg}\left(\left[e_{n}\right] \cdot\left[e_{m}\right]\right)=\operatorname{deg}\left[e_{n} \cdot e_{m}\right]=\operatorname{deg}\left[e_{n+m}\right]=n+m=\operatorname{deg}[f]+\operatorname{deg}[g]
$$

- (Multiplicative structure). Given any $\zeta \in S^{1}$ we have

$$
e_{n} e_{m}(\zeta)=e_{n}\left(\zeta^{m}\right)=\zeta^{m n}=\zeta^{n m}
$$

therefore $e_{n} e_{m}=e_{n m}$. From this we find that

$$
\operatorname{deg}([f] \circ[g])=\operatorname{deg}\left(\left[e_{n}\right] \circ\left[e_{m}\right]\right)=\operatorname{deg}\left[e_{n} \circ e_{m}\right]=\operatorname{deg}\left[e_{n m}\right]=n m=\operatorname{deg}[f] \cdot \operatorname{deg}[g]
$$

This finishes the proof that deg is an isomorphism of rings

$$
\begin{equation*}
\left[S^{1}, S^{1}\right] \simeq \mathbf{Z} \tag{দ}
\end{equation*}
$$

Corollary 15.5.9 (Automorphisms of $S^{1}$ ). If $f \in \operatorname{Aut}_{\text {Top }}\left(S^{1}\right)$ is an automorphism, then $\operatorname{deg} f= \pm 1$. As a consequence, either $f \sim_{h} \mathrm{id}_{S^{1}}$ or $f \sim_{h} \rho$, where $\rho$ represents the reflection by complex conjugation.
Proof. If $f$ is an automorphism, let $f^{-1}: S^{1} \rightarrow S^{1}$ be its inverse. Therefore $f f^{-1}=\operatorname{id}_{S^{1}}$, implying in

$$
1=\operatorname{deg} \operatorname{id}_{S^{1}}=\operatorname{deg}\left(f f^{-1}\right)=\operatorname{deg} f \cdot \operatorname{deg} f^{-1}
$$

which can only be the case if $\operatorname{deg} f=\operatorname{deg} f^{-1}= \pm 1$. Notice that the identity morphism is $\operatorname{id}_{S^{1}}=e_{1}$, therefore $\operatorname{deg} \operatorname{id}_{S^{1}}=1$. On the other hand, the reflection $\rho: S^{1} \rightarrow S^{1}$ maps $e^{2 \pi \mathrm{i} t} \mapsto e^{-2 \pi \mathrm{i} t}$-thus $\rho=e_{-1}$, implying in $\operatorname{deg} \rho=-1$. From the latter two considerations, the last proposition follows.
Example 15.5.10. A null-homotopic map $f: S^{1} \rightarrow S^{1}$ has $\operatorname{deg} f=0$, since it's homotopic to a constant map.
Corollary 15.5.11. The circle $S^{1}$ is not contractible.
Proof. Since deg id ${ }_{S^{1}}=1$, from our considerations from Example 15.5.10 we conclude that id ${ }_{S^{1}}$ isn't null-homotopic and therefore there exists no contraction of the circle. $\quad$,
Corollary 15.5.12. There exists no retraction $D^{2} \rightarrow S^{1}$.
Proof. Suppose, for the sake of contradiction, that there exists a retraction $r: D^{2} \rightarrow S^{1}$ such that $r \iota=\operatorname{id}_{S^{1}}$, where $\iota: S^{1} \hookrightarrow D^{2}$ is the canonical inclusion. Since $D^{2}$ is a contractible space, then $r$ is null-homotopic. This would imply that $r \iota$ was nullhomotopic and thus would be $\mathrm{id}_{\mathcal{S}^{1}}$, which contradicts Corollary 15.5.11. Therefore $r$ cannot exist.

## Applications of $\pi_{1}\left(S^{1}\right)$

Theorem 15.5.13 (Brouwer's fixed point). Every continuous map $D^{2} \rightarrow D^{2}$ has a fixed point.

Proof. Let $f: D^{2} \rightarrow D^{2}$ be any continuous map. Suppose, for the sake of contradiction, that $f$ admits no fixed point. We construct a map $r: D^{2} \rightarrow S^{1}$ as follows: for each $x \in D^{2}$, since $f x \neq x$, we can define a point $r x \in S^{1}$ given by $r x:=\frac{f x-x}{\|f x-x\|}$. Notice $r$ defines a retraction of the disk to the sphere, contradicting the result of Corollary 15.5.12, therefore $f$ must admit at least one fixed point.

Theorem 15.5.14 (Fundamental Theorem of Algebra). Every non-constant polynomial with coefficients in $\mathbf{C}$ has root in C .

Proof. Let $p(z)=z^{n}+a_{1} z^{n-1}+\cdots+a_{n}$ be any polynomial—note that the restriction to the case of monomials does not create any loss of generality. Suppose, for the sake of contradiction, that $p$ admits no roots in $\mathbf{C}$. Then for any real number $r \geqslant 0$
Finish this proof (look Hatcher's book)

## Chapter 16

## Covering Spaces

### 16.1 The van Kampen Theorem

## For the Fundamental Groupoid

Theorem 16.1.1 (van Kampen theorem for the fundamental groupoid). Let $X$ be a space, and $O$ be a connected open cover ${ }^{1}$ of $X$ that is closed under finite intersectionswhich is a subcategory of Top whose morphisms are inclusions. Considering the functor $\Pi_{1}:$ Top $\rightarrow$ Grpd, the fundamental groupoid $\Pi_{1} X$ is the colimit of the functor $\left.\Pi_{1}\right|_{o}$, that is:

$$
\Pi_{1} X \simeq \operatorname{colim}_{U \in O} \Pi_{1} U
$$

in the category Grpd.

Proof. We'll show that $\Pi_{1} \mid$ Let $\mathcal{G}$ be a groupoid and consider the constant functor $C: O \rightarrow \operatorname{Grpd}$ mapping $U \mapsto \mathcal{G}$ and $\iota \mapsto \mathrm{id}_{\mathcal{G}}$ for any object $U \in O$ and morphism $\iota$ of $O$. Let $\eta:\left.\Pi_{1}\right|_{o} \Rightarrow C$ be a natural transformation. The pair $(\mathcal{G}, \eta)$ forms a cocone over the functor $\left.\Pi_{1}\right|_{o}$ : indeed, given an object $U \in O$ there exists a morphism of groupoids $\eta_{U}: \Pi_{1} U \rightarrow C U=\mathcal{G}$ and from naturality, given any inclusion $\iota: U \hookrightarrow V$ in $O$ one has that

commutes in Grpd—showing that $\eta_{U}=\operatorname{id}_{\mathcal{G}} \eta_{U}=\eta_{V} \circ \Pi_{1} \iota$, hence compatibility $\Pi_{1}$ is satisfied, making $(G, \eta)$ a cocone.

Consider the collection $\left(i_{U}: \Pi_{1} U \rightarrow \Pi_{1} X\right)_{U \in O}$ of canonical inclusions of groupoids. To show the universal property of $\Pi_{1} X$ we must construct a unique morphism of

[^20]groupoids (that is, a functor between groupoids) $\chi: \Pi_{1} X \rightarrow \mathcal{G}$ such that

commutes in Grpd for all $U \in O$.
To that end, for every $x \in X$, if $x \in U$ let $\chi x:=\eta_{U} x$. For the morphisms of $\Pi_{1} X$, consider any path $f \in \operatorname{Path}_{X}(x, y)$ on $X$. If $\operatorname{im} f$ lies interely in some object $U \in O$, simply define $\chi[f]:=\eta_{U}[f]$-since $O$ is closed under finite intersections, if $\operatorname{im} f \subseteq U \cap V$ for some $U, V \in O$, then the image $\eta_{\bullet}[f]$ is independent of the choice of $U$ or $V$. Consider now the case where $f$ has an image not entirely contained in a single element of $O$, but multiple ones, say $\operatorname{im} f \subseteq \bigcup_{j=1}^{n} U_{j}$ for some finite collection of sets $U_{j} \in O$-since $O$ is closed under finite intersections, this, we shall define a corresponding collection of paths $\left(f_{j}: I \rightarrow U_{j}\right)_{j=1}^{n}$ such that
$$
f=f_{n} f_{n-1} \cdots f_{2} f_{1}
$$

With this collection in hands we may define $\chi[f]:=\eta_{U_{n}}\left[f_{n}\right] \cdots \eta_{U_{1}}\left[f_{1}\right]$.
We must ensure that this is well defined: let $g \in \operatorname{Path}_{X}(x, y)$ be another path and suppose there exists a homotopy $\varepsilon: f \Rightarrow g$. Take a decomposition $\left(g_{j}: I \rightarrow V_{j}\right)_{j=1}^{m}$ for sets $V_{j} \in O$. Consider a partition $\left(J_{j} \times S_{j}\right)_{j=1}^{\ell}$ of the square $I \times I$ such that im $\left.\varepsilon\right|_{J_{j} \times S_{j}} \subseteq U$ for some $U \in O$, and such that $\left(J_{j}\right)_{j=1}^{\ell}$ is a refinement for the decompositions of $f$ and $g$-that is, $\left.f\right|_{J_{j}}$ and $\left.g\right|_{J_{j}}$ are paths entirely contained in some set of $O$. In this way we see that $\varepsilon$ induces a collection of homotopies $\left(\varepsilon_{j}: f_{j} \Rightarrow g_{j}\right)_{j=1}^{\ell}$ proving that $\left[f_{j}\right]=\left[g_{j}\right]$ in some $\Pi_{1} U$, therefore $\chi[f]=\chi[g]$. Hence $\chi$ is a uniquely defined functor satisfying Eq. (16.1).

## For the Fundamental Group

Theorem 16.1.2 (van Kampen theorem for the fundamental group). Let $(X, x) \in \operatorname{Top}^{* /}$ be a path-connected space, and $O$ be a path-connected open cover of $X$ closed under finite intersections-and such that $x \in U$ for every $U \in O$, therefore $O$ is a subcategory of Top*/ whose morphisms are inclusions. Then the fundamental groupoid of $X$ is the colimit of the functor $\left.\pi_{1}\right|_{O}: O \rightarrow \mathrm{Grp}$, that is:

$$
\pi_{1}(X, x) \simeq \operatorname{colim}_{U \in O} \pi_{1}(U, x)
$$

We first prove a particular case of the classical van Kampen theorem and after generalize.

Lemma 16.1.3. The van Kampen theorem for the fundamental group holds when $O$ is finite.

Proof. Let $G$ be a group and $C: O \rightarrow$ Grp be the constant functor on $G$ and consider the cocone ( $G, \eta: \pi_{1} \mid O \Rightarrow C$ ) over the functor $\pi_{1} \mid O$. We'll construct a morphism of groups $\chi: \pi_{1}(X, x) \rightarrow G$. Recall that the inclusion functor $J: \pi_{1}(X, x) \rightarrow \Pi_{1} X$ is an equivalence of categories, since $\pi_{1}(X, x)$ is a skeleton of $\Pi_{1} X$ by Proposition 15.4.10. Define a quasi-inverse of $J$ as follows: consider a collection $\left(\gamma_{y}\right)_{y \in X}$ of paths $\gamma_{y} \in \operatorname{Path}_{X}(x, y)$ where $\operatorname{im} \gamma_{y} \subseteq U$ when $y \in U$ and $\gamma_{x}:=\operatorname{cons}_{x}$-this is possible because $O$ is closed under finite intersections-then define $F: \Pi_{1} X \rightarrow \pi_{1}(X, x)$ by mapping $f: a \rightarrow b$ to $F f:=\gamma_{b} f \gamma_{a}^{-1}: x \rightarrow x$.

Notice that the quasi-inverse functors $J$ and $F$ induce, for each $U \in O$, a corresponding pair of quasi-inverse functors

$$
F_{U}: \Pi_{1} U \rightleftarrows \pi_{1}(U, x): J_{U} .
$$

Then we can construct a cocone $\left(G, \delta:\left.\Pi_{1}\right|_{O} \Rightarrow C\right)$ over the functor $\left.\Pi_{1}\right|_{O}$, where $\delta_{U}:=$ $\eta_{U} F_{U}: \Pi_{1} U \rightarrow G$. By means of Theorem 16.1.1 there exists a unique morphism of groupoids $\xi: \Pi_{1} X \rightarrow G$ (where $G$ is interpreted as a groupoid with a single object) such that

commutes in Grpd for every $U \in O$. Define $\chi:=\xi J: \pi_{1}(X, x) \rightarrow G$, and notice that since $\eta_{U} F_{U}=\xi i_{U}$ we can precompose with $J_{U}: \pi_{1}(U, x) \rightarrow \Pi_{1} U$ and use that $F_{U} J_{U}=\operatorname{id}_{\pi_{1}(U, x)}$ to obtain that $\eta_{U}=\xi i_{U} J_{U}$. Notice however that given any $[g] \in \pi_{1}(U, x)$ one has $i_{U} J_{U}[g]=[g] \in \Pi_{1} X$ while $J j_{U}[g]=[g] \in \Pi_{1} X$ again-for the canonical inclusion $j_{U}: \pi_{1}(U, x) \hookrightarrow \pi_{1}(X, x)$ —therefore $i_{U} J U=J j_{U}$. This proves that $\eta_{U}=\xi J j_{U}=\chi j_{U}$ for every $U \in O$, that is

commutes in Grp, which proves the universal property for the colimit $\pi_{1}(X, x)$.

## Proof of the Classical van Kampen Theorem

Let $O$ be a path-connected open cover of $X$ closed under intersections and composed of neighbourhoods of the chosen base-point $x$. Let $\mathfrak{F} \subseteq 2^{O}$ be the category whose objects are the finite subsets of $O$ of the cover that is closed under finite intersections, and morphisms are inclusions. Given any such subset $C \in \mathscr{F}$, we know from Lemma 16.1.3 that the space $U_{C}:=\bigcup_{U \in C} U$ satisfies

$$
\begin{equation*}
\pi_{1}\left(U_{C}, x\right) \simeq \operatorname{colim}_{U \in C} \pi_{1}(U, x) \tag{16.2}
\end{equation*}
$$

- Let's prove that the colimit of the functor $\left.\pi_{1}\right|_{\mathfrak{F}}: \mathfrak{F} \rightarrow$ Grp-which maps each $C \in \mathscr{F}$ to the group $\pi_{1}\left(U_{C}, x\right)$-is the fundamental group $\pi_{1}(X, x)$. Given any group $G$ and a its corresponding constant functor $C_{G}: \mathfrak{F} \rightarrow$ Grp with $C_{G} C:=G$, let $\eta:\left.\pi_{1}\right|_{\tilde{F}} \Rightarrow C_{G}$ be a natural transformation. The pair $(G, \eta)$ is then a cocone over the functor $\left.\pi_{1}\right|_{\mathfrak{F}}$.
We'll construct a unique morphism of groups $\chi: \pi_{1}(X, x) \rightarrow G$ satisfying the coherence of the cocones using the same technique from Theorem 16.1.1. If $f: x \rightarrow x$ is a loop contained entirely in a set $U_{C} \subseteq X$ for some $C \in \mathscr{F} F$, we simply map $\chi[f]:=\eta_{C}[f]$. If on the other hand $f$ is not entirely contained in a single set, say that $f$ is contained in the union $\bigcup_{j=1}^{n} U_{j}$ for sets $U_{j} \in C$ and define a collection of decompositions of $f$, namely $\left(f_{j}: I \rightarrow U_{j}\right)_{j=1}^{n}$, for which $f$ is the result of the concatenation of paths. From the same argument as before, merely map $\chi[f]:=\eta_{U_{n}}\left[f_{n}\right] \cdots \eta_{U_{1}}\left[f_{1}\right]$, which is well defined and unique ${ }^{2}$. Therefore one has

$$
\begin{equation*}
\operatorname{colim}_{C \in \mathfrak{F}} \pi_{1}\left(U_{C}, x\right) \simeq \pi_{1}(X, x) \tag{16.3}
\end{equation*}
$$

- For the final part of the proof, we shall prove that the colimits of the functors $\left.\pi_{1}\right|_{o}$ and $\left.\pi_{1}\right|_{\mathfrak{F}}$ agree so that the van Kampen theorem is true. Recalling Eq. (16.2), one has

$$
\begin{aligned}
\operatorname{colim}_{C \in \mathfrak{F}} \pi_{1}\left(U_{C}, x\right) & \simeq \operatorname{colim}_{\mathcal{C} \in \mathscr{F}}\left(\operatorname{colim}_{U \in C} \pi_{1}(U, x)\right) \\
& \simeq \operatorname{colim}_{(O, \mathscr{F})} \pi_{1}(U, x)
\end{aligned}
$$

where $(O, \mathfrak{F})$ is the category whose objects are pairs $(U, C) \in C \times \mathfrak{F}$, and morphisms are paired inclusions-also $\left.\pi_{1}(-, x)\right|_{(O, \mathscr{F})}:(O, \mathscr{F}) \rightarrow$ Grp is defined to map $(U, C)$ to $\pi_{1}(U, x)$. Notice that the functors $\left.\pi_{1}(-, x)\right|_{O}$ and $\left.\pi_{1}(-, x)\right|_{(O, \mathfrak{F})}$ factor as


Where $\iota U:=(U,\{U\})$ and $p(U, C):=U$. Therefore one has an isomorphism

$$
\begin{equation*}
\operatorname{colim}_{U \in O} \pi_{1}(U, x) \simeq \operatorname{colim}_{(U, C) \in(O, \overparen{F})} \pi_{1}(U, x) \tag{16.4}
\end{equation*}
$$

Therefore, by Eqs. (16.3) and (16.4) we have

$$
\operatorname{colim}_{U \in O} \pi_{1}(U, x) \simeq \pi_{1}(X, x)
$$

as wanted.

[^21]
### 16.2 Covering Spaces

## Initial Constructions

Definition 16.2.1 (Covering space). A surjective continuous map $p: E \rightarrow B$ is said to be a covering space over $B$ if for every point $x \in B$, there exists a neighbourhood $U \subseteq B$ of $x$ that is evenly covered by $p$, that is: there exists a bundle isomorphism over $B$ between the pullback ${ }^{3}$ of $p$ over $U$ and a product bundle $\pi: U \times E_{x} \rightarrow U$ with discrete ${ }^{4}$ fibre $E_{x}=p^{-1} x$. Diagrammatically, one has that the diagram

commutes in Top, where the topological isomorphism $U \times E_{x} \xrightarrow{\simeq} p^{-1} U$ is explicitly given by $(u, e) \mapsto e$. Said concisely, a covering space is a locally trivial bundle with discrete fibre.

Yet another equivalent way of defining a covering space goes as follows. $B$ is evenly covered by $p$ if there exists an open cover $\mathcal{U}$ of $B$ such that any $U \in \mathcal{U}$ has a preimage $p^{-1} U=\coprod_{\alpha} V_{\alpha}$ where $V_{\alpha}$ is an open set of $E$, and the restriction $\left.p\right|_{V_{\alpha}}: V_{\alpha} \xrightarrow{\approx} U$ is a topological isomorphism.

Corollary 16.2.2. A covering space is a local topological isomorphism.
Proof. Given a covering space $p: E \rightarrow B$ and $x \in E$, we can consider the neighbourhood $U \subseteq B$ of $p x$ that is evenly covered by $p$. Then from our last definition we know that $U$ is isomorphic to a disjoint union of open sets of $E$. Let $V \subseteq p^{-1} U$ be one such open set which is also a neighbourhood of $x$. Then the restriction $\left.p\right|_{V}: V \rightarrow p(V)=U$ is an isomorphism, showing that $p$ is indeed a local isomorphism.

Notation 16.2.3 (Sheets). Let $p: E \rightarrow B$ be a covering space and $x \in B$ be any point. Consider a neighbourhood $U \subseteq B$ of $x$ such that $p$ admits a trivialisation $\phi: p^{-1} U \xrightarrow{\leftrightharpoons}$ $U \times E_{x}$. Since $E_{x}$ is discrete, we have an isomorphism $U \times E_{x} \simeq \coprod_{e \in E_{x}} U \times e$. Moreover, clearly $U \simeq U \times e$ for any $e \in E_{x}$. We shall define a sheet over $U$ to be an open set

$$
\begin{equation*}
U_{e}:=\phi^{-1}(U \times\{e\}) \subseteq E \tag{16.5}
\end{equation*}
$$

The motivation comes from the fact that $U_{e} \simeq U \times e \simeq U$.
Proposition 16.2.4 (Covering projections are open). Given a covering space $p: E \rightarrow B$, the projection $p$ is open.

[^22]Proof. Consider the open cover $\left(U_{x}\right)_{x \in B}$ of $B-w h e r e U_{x}$ is a neighbourhood of $x$ such that there exists a topological isomorphism $p^{-1} U_{x} \simeq U_{x} \times \operatorname{Disc} E_{x}$, for each $x \in B$. Under the product topology, projections are open maps (see Lemma 13.3.5), therefore $\left.p\right|_{p^{-1} U_{x}}$ is an open map for each $x \in B$. Given any open set $V \subseteq E$, notice that $V=\bigcup_{x \in B}\left(W \cap p^{-1} U_{x}\right)$, therefore

$$
p V=p\left(\bigcup_{x \in B}\left(W \cap p^{-1} U_{x}\right)\right)=\bigcup_{x \in B} p\left(W \cap p^{-1} U_{x}\right)
$$

is the union of open sets, thus open-which shows that $p$ is an open map.
Lemma 16.2.5 (Fibre-wise diagonal of covering space is open and closed). Let $p: E \rightarrow B$ be a covering space, and $E \times_{B} E$ be the pullback of $p$ with itself. Then the diagonal of $E$ with respect to the fibre product over $B$, namely

$$
\Delta_{B} E=\left\{(e, e) \in E \times_{B} E: e \in E\right\}
$$

is an open and closed set in $E \times{ }_{B} E$.
Proof. First we prove that $\Delta_{B} E$ is open. Let $e \in E$ be any point and $U_{p e} \subseteq B$ be a neighbourhood of $p e$ together with an isomorphism

$$
p^{-1} U_{p e} \simeq U_{p e} \times E_{p e}=U_{p e} \times\{e\} .
$$

Since $p^{-1} U_{p e}$ is open, it follows that $U_{p e} \times\{e\} \hookrightarrow E$ is open. Then the set $\left(E \times_{B} E\right) \cap$ $\left(U_{p e} \times U_{p e}\right)$ is an open neighbourhood of $(e, e)$ in $E \times_{B} E$. This shows that $\Delta_{B} E$ is open.

To show that $\Delta_{B} E$ is closed, let $(x, y) \in E \times_{B} E$ with $x \neq y$. Let $U \subseteq B$ be a neighbourhood of $p x=p y$ that is evenly covered, and consider the induced sheets $U_{x}, U_{y} \subseteq E$ for $x$ and $y$, respectively. From the assumption that $x$ and $y$ are distinct points, we have an empty intersection $U_{x} \cap U_{y}=\varnothing$-which shows that $U_{x} \times U_{y}$ is disjoint from $\Delta_{B} E$. Therefore $\left(E \times_{B} E\right) \cap\left(U_{x} \times U_{y}\right)$ is a neighbourhood for $(x, y) \in E \times_{B} E$ outside of $\Delta_{B} E$, showing that $\Delta_{B} E$ is closed.

### 16.3 Lifting Properties

Definition 16.3.1 (Lift). Let $f: Y \rightarrow X$ and $g: Z \rightarrow X$ be morphisms. We define a lift of $g$ along $f$ to be a morphism $\widehat{g}: Z \rightarrow Y$ such that the following diagram commutes:


Proposition 16.3.2 (Homotopy lift via covering space). Let $p: E \rightarrow B$ be a covering space and consider a homotopy $\eta: X \times I \rightarrow B$. Let $\widehat{\eta}_{0}: X \rightarrow E$ be a lift of $\eta(-, 0)$ along $p$.

Then there exists a unique lift $\widehat{\eta}: X \times I \rightarrow E$ of $\eta$ along $p$ which extends $\widehat{\eta}_{0}$. That is, the following diagram commutes:


Proof. We shall prove this proposition in three acts, first we shall construct a lift with respect to a neighbourhood of any given point $x \in X$, later the uniqueness of such lift, and finally how all this pieces together to form the required lift.
(i) Let $x \in X$ be any point and take, for each $(x, t) \in x \times I$, an evenly covered neighbourhood $U_{t} \subseteq B$ of $\eta(x, t)$. Now we can choose a neighbourhood $N_{t} \times$ $\left(a_{t}, b_{t}\right)$ of $(x, t)$ such that $\eta\left(N_{t} \times\left(a_{t}, b_{t}\right)\right) \subseteq U_{t}$ is again evenly covered by $p$.
Since $x \times I$ is compact, let $0=t_{0}<t_{1}<\cdots<t_{n}=1$ be a partition of $I$ such that $\left\{N_{t_{j}} \times\left[t_{j}, t_{j+1}\right]\right\}_{j=0}^{n-1}$ covers $x \times I$-and also satisfies the requirement that $\eta\left(N_{t_{j}} \times\left[t_{j}, t_{j+1}\right]\right)$ is evenly covered by $p$ as in the previous construction. Define the neighbourhood $N:=\bigcap_{j=0}^{n} N_{t_{j}}$ of $x$, so that $\eta\left(N \times\left[t_{j}, t_{j+1}\right]\right)$ is still evenly covered by $p$ for any $j$.
We proceed by induction assuming that, for some $j$, the lift $\widehat{\eta}$ has been constructed on $N \times\left[0, t_{j}\right]$ and which extends the given starting lift $\widehat{\eta}_{0}$. Let $U_{j} \subseteq B$ be an evenly covered set containing $\eta\left(N \times\left[t_{j}, t_{j+1}\right]\right)$, and take $V_{j} \subseteq E$ to be an open set projecting isomorphically onto $U_{j}$ via $p$ and with $\widehat{\eta}(x, t) \in V_{j}$. Define

$$
N^{\prime} \times t_{j}:=\left(N \times t_{j}\right) \cap\left(\left.\widehat{\eta}\right|_{N \times t_{j}}\right)^{-1} V_{j},
$$

which is again a neighbourhood of $\left(x, t_{j}\right)$. From this construction we find that $\widehat{\eta}\left(N^{\prime} \times t_{j}\right) \subseteq V_{j}$. Now we define $\widehat{\eta}$ on $N^{\prime} \times\left[t_{j}, t_{j+1}\right]$ to be the composition of $\eta$ with $p^{-1}: U_{j} \rightarrow V_{j}$, that is:


This procedure is to be done a finite number of iterations, so that in the end we obtain a lift $\widehat{\eta}: M \times I \rightarrow E$ for some neighbourhood $M \subseteq X$ of $x$.
(ii) To prove the uniqueness of the lifting $\widehat{\eta}$ we'll work simply on the case of a single point $x$, so that uniqueness for the total set $X \times I$ will follow from the uniqueness of the lift on segments $x \times I$ for each $x \in X$. Let $\widehat{\eta}$ and $\widehat{\eta}^{\prime}$ be two lifts of $\eta: x \times I \rightarrow B$ with $\widehat{\eta}(0)=\widehat{\eta}^{\prime}(0)$. From the same argument as before, take a partition $0=t_{0}<t_{1}<\cdots<t_{n}=1$ of $I$ such that $\eta\left(\left[t_{j}, t_{j+1}\right]\right)$ is contained in an evenly covered neighbourhood $U_{j} \subseteq B$ by $p$. Using induction, we may assume that $\left.\widehat{\eta}\right|_{\left[0, t_{j}\right]}=\widehat{\eta}_{\left[0, t_{j}\right]}$ for some $j$. Using the fact that $\left[t_{j}, t_{j+1}\right]$ is a connected set,
the image $\widehat{\eta}\left(\left[t_{j}, t_{j+1}\right]\right) \subseteq E$ will also be connected. Therefore there exists a set $V_{j} \subseteq E$ containing $\widehat{\eta}\left(\left[t_{j}, t_{j+1}\right]\right)$, and with $V_{j} \simeq U_{j}$ via $p$. Since $\widehat{\eta} t_{j}=\widehat{\eta}^{\prime} t_{j}$ then by connectedness $\widehat{\eta}^{\prime}\left(\left[t_{j}, t_{j+1}\right]\right) \subseteq V_{j}$. The lift condition implies in $p \widehat{\eta}=p \widehat{\eta}^{\prime}$, however we also know that $p$ is injective in $V_{j}$-therefore it follows that $\widehat{\eta}_{\left[t_{j}, t_{j+1}\right]}=\left.\widehat{\eta}^{\prime}\right|_{\left[t_{j}, t_{j+1}\right]}$. This procedure is continued for finitely many steps and shows that $\widehat{\eta}=\widehat{\eta}^{\prime}$.
(iii) Notice that, by choosing a point $x \in X$, we obtain a lift piece of the form $M \times I \rightarrow E$ where $M \subseteq X$ is a neighbourhood of $x$. Notice however that in the overlap of such neighbourhoods, when one considers the whole range of points contained in $X$, we must have an agreement of the lift pieces. Therefore, such neighbourhoods may be glued to create a well defined lift $\widehat{\eta}: X \times I \rightarrow E$. Moreover, using the proof of the last item we know that each segment $\left.\widehat{\eta}\right|_{x \times I}$ is unique for any $x \in X$, thus $\widehat{\eta}$ itself is unique.

Corollary 16.3.3. Let $p: E \rightarrow B$ be a covering space. The following are lifting properties associated to $p$ :
(a) For each path $\gamma: I \rightarrow B$ starting at $x_{0} \in B$ and each $e_{0} \in p^{-1} x_{0}$ there exists a unique lift $\widehat{\gamma}: I \rightarrow E$ which starts at $e_{0}$.
(b) For each homotopy $\eta: I \times I \rightarrow B$ of paths starting at $x_{0}$ and each $e_{0} \in p^{-1} x_{0}$ there is a unique lifted homotopy $\widehat{\eta}: I \times I \rightarrow E$ of paths starting at $e_{0}$.
(c) The lift of a constant path is constant.
(d) Every homotopy between paths lifts to a homotopy of paths ${ }^{5}$.

Proposition 16.3.4. Given a covering $p: E \rightarrow B$, the induced morphism of groups

$$
p_{*}: \pi_{1}(E, e) \mapsto \pi_{1}(B, p e)
$$

is a monomorphism for any $e \in E$.
Proof. Let $[\widehat{f}]$ be a kernel element of $p_{*}$, so that $p_{*}[\widehat{f}]=[f]=\left[\right.$ cons $\left._{p e}\right]$. Thus there exists a homotopy $\eta: f \Rightarrow$ cons $_{p e}$ and via Proposition 16.3 .2 we obtain a lifted homotopy of loops $\widehat{\eta}: \widehat{f} \Rightarrow$ cons $_{e}$, showing that $[\widehat{f}]=\left[\right.$ cons $\left._{e}\right]$. Therefore $p$ is indeed a monomorphism of groups.

Proposition 16.3.5. Let $E$ and $B$ be path-connected spaces. The number of sheets of a pointed covering space $p:(E, e) \rightarrow(B, b)$ is equal to the index of the subgroup $p_{*} \pi_{1}(E, e)$ in $\pi_{1}(B, b)$.

Proof. Define the notation $G:=p_{*} \pi_{1}(E, e)$. Let $\gamma \in \Omega(B, b)$ be a loop and consider its lifted loop $\widehat{\gamma} \in \Omega(E, e)$.

## Continue

[^23]Theorem 16.3.6 (Lifting out of connected space). Let $p: E \rightarrow B$ be a covering space, and $f: Y \rightarrow X$ be a continuous map, where $Y$ is a connected space. Consider two lifts of $f$ along $p$ : continuous maps $\widehat{f}_{1}, \widehat{f}_{2}: Y \rightrightarrows E$ such that the triangle

commutes in Top. If there exists $y \in Y$ such that $\widehat{f}_{1} y=\widehat{f}_{2} y$, then the lifts agree everywhere $\widehat{f}_{1}=\widehat{f}_{2}$.

Proof. Consider the pullback $E \times_{B} E$ of $p$ with itself and consider the uniquely defined $\operatorname{morphism}\left(\widehat{f}_{1}, \widehat{f}_{2}\right): Y \rightarrow E \times{ }_{B} E$ making the diagram

commute. Define $\Delta_{B} E \subseteq E \times_{B} E$ for the diagonal of $E$ with respect to the fibre product. Using Lemma 16.2 .5 we know that $\Delta_{B} E$ is both open and closed, thus $V:=\left(\widehat{f}_{1}, \widehat{f}_{2}\right)^{-1}\left(\Delta_{B} E\right) \subseteq Y$ is both open and closed in $Y$, which is a non-empty set since by hypothesis $\widehat{f}_{1}$ and $\widehat{f}_{2}$ agree at least in one point of $Y$. Since $V$ is closed, then $Y \backslash V$ is open in $Y$ and certainly disjoint from $V$. Since their union is the whole space $Y$, by the hypothesis that $Y$ is connected, it must be the case that $V=Y$. Therefore $\widehat{f}_{1}=\widehat{f}_{2}$ as wanted.

Proposition 16.3.7. Let $q: E \rightarrow B \times I$ be a locally trivial covering space with fibre $F$. Then $B$ admits an open cover $\mathcal{U}$ for which $q$ is trivial over $U \times I$ for each $U \in \mathcal{U}$.

## Prove when needed

### 16.4 Coverings and G-Actions

Definition 16.4.1 (Properly discontinuous action). Given a discrete topological group $G$, a left action $G \times E \rightarrow E$ is said to be properly discontinuous if for each pair $(g, x) \in G \times E$, where $g \neq e$, there exists a neighbourhood $U \subseteq E$ of $x$ such that

$$
U \cap g U=\varnothing .
$$

In particular, every properly discontinuous action is free.
Definition 16.4.2 ( $G$-principal covering space). Let $G$ be a discrete topological group. A left G-principal covering space is a covering $p: E \rightarrow B$ together with a properly discontinuous left action $G \circlearrowright E$ for which $p(g x)=p x$ for every pair $(g, x) \in G \times E$, and such that the induced action on the fibres is transitive.

Proposition 16.4.3. If $G \circlearrowright X$ is a properly discontinuous action of a discrete group $G$ on a space $X$, the canonical projection $q: X \rightarrow X / G$ is a $G$-principal covering space.

Proof. First of all, it is clear that $q(g x)=q x \in X / G$ for any pair $(g, x) \in G \times X$. Let $[x] \in$ $X / G$ be any point, and $V \subseteq X$ be a neighbourhood of $x$ such that $V \cap g V \neq \varnothing$ implies $g=e$, and define $U:=q V$. Notice that since $G$ acts by topological isomorphisms, one has $q^{-1} U=\bigcup_{g \in G} g V$-now, since each $g V$ is open, it follows that $q^{-1} U \subseteq X$ is open. Since $X / G$ has the quotient topology, then $U \subseteq X / G$ is open.

Consider $q^{-1} U=\coprod_{g \in G} V_{g}$, where $V_{g}:=g V$. Given any $g \in G$, suppose there exists $h \in G$ such that $V_{g} \cap V_{h} \neq \varnothing$-then for any $x^{\prime} \in V_{g} \cap V_{h}$, one has $h^{-1} x^{\prime} \in V_{g} \cap V$, and from construction this implies in $h^{-1} g=e$, thus $h=g$.

It remains for us to show a trivialisation for $q$-we shall prove that $\left.q\right|_{V_{g}}: V_{g} \rightarrow U$ is an isomorphism. Since quotient maps are open, it suffices to show that $\left.q\right|_{V_{g}}$ is bijective. Let $\left[x^{\prime}\right] \in U$ be any point, and take $x^{\prime \prime} \in V$ such that $q x^{\prime \prime}=\left[x^{\prime}\right]$, then in particular $g x^{\prime \prime} \in g V$ and $q\left(g x^{\prime \prime}\right)=\left[g x^{\prime \prime}\right]=\left[x^{\prime}\right]$-thus $\left.q\right|_{V_{g}}$ is surjective. For injectivity, let $x^{\prime}, x^{\prime \prime} \in V$ be a pair of points such that $q\left(g x^{\prime}\right)=q\left(g x^{\prime \prime}\right)$, then $\left[x^{\prime}\right]=\left[x^{\prime \prime}\right]$ in $X / G$, which implies in the existence of a point $h \in G$ such that $x^{\prime}=h x^{\prime \prime}$, then $x^{\prime \prime} \in V \cap V_{h}$-therefore $h=e$ and hence $x^{\prime}=x^{\prime \prime}$, thus in particular $g x^{\prime}=g x^{\prime \prime}$ as wanted.

Definition 16.4.4 (Deck transformations). Given a covering space $p: E \rightarrow B$, we define a group of automorphisms $\operatorname{Aut}(p)$ of the cover $p$ to be composed of topological isomorphisms $\alpha: E \xrightarrow{\simeq} E$ such that $p \alpha=p$-such maps are called deck transformations of the covering $p$.

Example 16.4.5 (Translations). Given a left $G$-principal covering $p: E \rightarrow B$, the left translation of $E$ by $g \in E$, the isomorphism $\ell_{g}: E \xrightarrow{\leftrightharpoons} E$ mapping $e \mapsto g e$, is a deck transformation of $p$. The collection of such maps $\left(\ell_{g}\right)_{g \in G}$ define a morphism of groups

$$
\ell: G \longrightarrow \operatorname{Aut}(p)
$$

Assume that $E$ is a connected space. Let $x \in B$ be any point and consider any deck transformation $\alpha \in \operatorname{Aut}(p)$. From the fact that $\alpha$ is bijective and $p \alpha=p$, it acts as a permutation on the fibre $p^{-1} x$. Since $p$ is a $G$-principal covering, then $G$ acts transitively on the fibres: hence given any two points $e, e^{\prime} \in p^{-1} x$, since $\alpha e \in p^{-1}(p e)$, it follows that there exists $g \in G$ such that $\alpha e^{\prime}=g(\alpha e)$. Therefore $\alpha$, using the connectedness of $E$, is uniquely defined by its image under a single point. This shows that, under these assumptions, $\ell$ is an isomorphism $G \simeq \operatorname{Aut}(p)$.

Proposition 16.4.6. Let $p: E \rightarrow B$ be a covering space. The following are properties concerning the automorphism group of $p$ :
(a) If $E$ is a connected space, then $\operatorname{Aut}(p)$ has a properly discontinuous action on $E$.
(b) If $B$ is locally path connected and $H \subseteq \operatorname{Aut}(p)$ is a subgroup, then the induced map $q: E / H \rightarrow B$ is a covering.

Proof. (a) Consider any point $e \in E$ and deck transformation $g \in \operatorname{Aut}(p)$. Since $p$ is a covering, choose $U \subseteq B$ to be an evenly covered neighbourhood of $p e$ and let $U_{e}$ be a sheet over $U$ with $e \in U_{e}$. Suppose there exists a point $e^{\prime} \in U_{e} \cap g U_{e}$ and notice that, since $p g=p$, we have $p e^{\prime}=p\left(g^{-1} e^{\prime}\right)$.

$$
\begin{aligned}
& \text { Continue proof of the properties of the automorphism group of covering } \\
& \text { spaces. }
\end{aligned}
$$

Proposition 16.4.7. Let $E$ be a simply connected space. If $G \circlearrowright E$ is a properly discontinuous action of a discrete topological group $G$, then the action induces an isomorphism of groups

$$
\pi_{1}(E / G) \simeq G .
$$

Proof. Let $[x] \in E / G$ be any base point, and choose $x \in q^{-1}[x]$-where $q: E \rightarrow E / G$ is the canonical projection. Define a map $\psi_{x}: \pi_{1}(E / G,[x]) \rightarrow G$ such that $\psi_{x}[\alpha]=g$ if and only if $\widehat{\alpha}_{x}(1)=g x$-where $\widehat{\alpha}_{x}: I \rightarrow E$ is the lift of $\alpha$ over $q$. We now show that $\psi_{x}$ is the required isomorphism of groups:

- (Well defined) Suppose that $[\alpha]=[\beta]$ and that $\psi_{x}[\alpha]=g$ while $\psi_{x}[\beta]=h$. This means that $\widehat{\alpha}_{x}(1)=g x$ and $\widehat{\beta}_{x}(1)=h x$, but since $\widehat{\alpha}_{x}=\widehat{\beta}_{x}$, then $g x=h x$ and hence $h^{-1} g x=x$. Since $G$ has a properly discontinuous action, there exists a neighbourhood $U \subseteq E$ of $x$ such that $U \cap\left(h^{-1} g\right) U$ is non-empty, thus it is necessarily the case that $h^{-1} g=e$, thus $h=g$. This proves that $\psi_{x}[\alpha]=\psi_{x}[\beta]$.
- (Injective) Suppose $\psi_{x}[\alpha]=\psi_{x}[\beta]$, then $\widehat{\alpha}_{x}(1)=g x=\widehat{\beta}_{x}(1)$ for some $g \in G$. This

- (Surjective) From the fact that $\pi_{1}(E / G,[x])$ acts transitively on the fibres of $q$, it follows that $\psi_{x}$ is surjective.
- (Group morphism) Let $\psi_{x}[\alpha]=g$ and $\psi_{x}[\beta]=h$ and define $k:=\psi_{x}[\beta \cdot \alpha]$, so that

$$
k x=(\widehat{\beta \cdot \alpha})_{x}(1)=\left(\widehat{\beta}_{\widehat{\alpha}_{x}(1)} \cdot \widehat{\alpha}_{x}\right)(1)=\widehat{\beta}_{g x}(1)=g \widehat{\beta}_{x}(1)=g h x
$$

therefore $k=g h$.

### 16.5 Fibre Transport

Definition 16.5.1 (Homotopy lifting property). A continuous map $p: E \rightarrow B$ is said to have the homotopy lifting property (HLP) for a given space $X$ if: for any homotopy
$\eta: X \times I \rightarrow B$ and continuous map $a: X \rightarrow E$ such that $p a x=\eta(x, 0)$ exists a homotopy $\delta: X \times I \rightarrow E$ such that the diagram

commutes in Top-the map $i_{0}: X \hookrightarrow X \times I$ is defined as $x \mapsto(x, 0)$
Check proof of these last fibration stuff-tom Dieck's book

Definition 16.5.2 (Fibration). A continuous map $p: E \rightarrow B$ is said to be a fibration if it satisfies the homotopy lifting property for every space.

Theorem 16.5.3. A covering space is a fibration.
Proposition 16.5.4 (Unique path lifting). Let $p: E \rightarrow B$ be a covering space, and $\gamma: I \rightarrow B$ be a path with initial point $\gamma 0:=p e \in B$. Then there exists a unique lifting $I \rightarrow E$ of $\gamma$ that begins in $e$. Furthermore, any two paths $u, v: I \rightrightarrows E$ with $u 0=v 0$ are homotopic if and only if their images are homotopic in $B$.

Lemma 16.5.5. Let $p: E \rightarrow B$ be a covering space and define $p e_{0}:=b_{0}$. Let $Y$ be a path connected and locally path connected space, and $f: Y \rightarrow B$ be a continuous map with $f y_{0}=b_{0}$. There exists a lift $Y \rightarrow E$ of $f$ over $p$ such that $y_{0} \mapsto e_{0}$ if and only if

$$
f_{*} \pi_{1}\left(Y, y_{0}\right) \subseteq p_{*} \pi_{1}\left(E, e_{0}\right)
$$

If such a lift exists, then it's unique.
See Munkres, page 478

## Chapter 17

## Cellular Structures

### 17.1 Compactly Generated Spaces

## Weak-Hausdorff Spaces

Definition 17.1.1 (Weak-Hausdorff space). A space $X$ is said to be weak-Hausdorff if for every compact Hausdorff space $K$ and continuous map $g: K \rightarrow X$ one has that $g(K)$ is closed in X.

Corollary 17.1.2. A weak-Hausdorff space is $\mathrm{T}_{1}$.
Proof. Let $X$ be a weak-Hausdorff space and $x \in X$, we shall take advantage of the fact that $X$ is $\mathrm{T}_{1}$ if and only if $\{x\}$ is closed in $X$. Let $K$ be a compact set and cons $x_{x}: K \rightarrow X$ be the constant map $k \mapsto x$, which is continuous. Since $g K=\{x\}$, and $g K$ is closed by hypothesis, the property follows.

Lemma 17.1.3. If $X$ is weak-Hausdorff, $K$ is a compact Hausdorff space, and $g: K \rightarrow X$ is a morphism, then $g K$ is a compact Hausdorff subspace of $X$.

Proof. We already know from Proposition 14.2.2 that $g K$ is compact, thus we shall only prove that $g K$ is Hausdorff. Let $x, y \in g K$ be any two distinct points. Using Corollary 17.1.2 and Proposition 14.6 .2 we can find disjoint open sets $U, V \subseteq K$ such that $g^{-1} x \subseteq U$ and $g^{-1} y \subseteq V$. Then the sets $K \backslash U$ and $K \backslash V$ are closed in $K$, thus compact (see Proposition 14.2.5). Therefore the sets $g(K \backslash U)$ and $g(K \backslash V)$ are both closed in $X$. From this we know that $g K \backslash g(K \backslash U)$ and $g K \backslash g(K \backslash V)$ are both open in $g K$, both of which are disjoint and contain $x$ and $y$, respectively-which shows that $X$ is Hausdorff.

Definition 17.1.4 (Compactly closed). A subset $A$ of $X$ is said to be compactly closed if, for every compact space $K$ and morphism $g: K \rightarrow X$, the preimage $g^{-1} A$ is closed in $K$.

Proposition 17.1.5. If $X$ is a weak-Hausdorff space, a subset $A \subseteq X$ is compactly closed if and only if the intersection of $A$ with each compact subset of $X$ is closed.

Proof. - $(\Rightarrow)$ Suppose $A$ is compactly closed, and let $K \subseteq X$ be any compact subspace of $X$. If we consider the inclusion morphism $t: K \hookrightarrow X$, we have that $\iota^{-1} A=K \cap A$ is closed.

- $(\Leftarrow)$ Assume the latter property, and let $K$ be a compact space together with a morphism $g: K \rightarrow X$. Since $g K$ is compact in $X$ it follows from assumption that $A \cap g K$ is closed in $X$. Since $g$ is continuous, then $g^{-1}(A \cap g K)=g^{-1} A$ is closed in $K$.


## $k$-Spaces

Definition 17.1.6 ( $k$-space). A space $X$ is said to be a $k$-space if every compactly closed subspace of $X$ is closed. The full-subcategory of Top whose objects are $k$-spaces will be denoted by $k$-Top.

Lemma 17.1.7 ( $k$-ification). Given a topological space $(X, \tau)$, we can transform $X$ into a $k$-space by creating a topology $\tau_{k}$ where $C$ is closed in $\left(X, \tau_{k}\right)$ if and only if $C$ is compactly closed in $(X, \tau)$. We shall shortly denote the $k$-space $\left(X, \tau_{k}\right)$ by $k X$.

Definition 17.1.8 ( $k$-ification functor). We define the $k$-ification functor

$$
k: \text { Top } \longrightarrow k \text {-Top }
$$

to be the functor mapping topological spaces $X$ to its $k$-ified space $k X$, and $k f:=f$ for every topological morphism $f$.

Lemma 17.1.9. If $X$ is a weak-Hausdorff space, then $k X$ is also a weak-Hausdorff $k$-space.

Notation 17.1.10 (Products). In what follows, we shall denote by $X \times_{c} Y$ the cartesian product of spaces $X$ and $Y$, which shall be endowed with the usual product topology. Moreover, from now on we shall reserve the notation

$$
X \times Y:=k\left(X \times_{c} Y\right)
$$

which may seem odd, but is a convention used throughout the literature and we'll adopt here.

Proposition 17.1.11 (Quotients). The quotient of a $k$-space is a $k$-space.
Proof. Let $X$ be a $k$-space and $q: X \rightarrow Y$ be a quotient map-we want to show that $Y$ is a $k$-space. Let $A \subseteq Y$ be a compactly closed subset. We'll prove that $q^{-1} A$ is compactly closed in $X$, thus closed, yielding the conclusion that $A$ is closed in $Y$. Given any compact space $K$ and a morphism $\phi: K \rightarrow X$. Then the map $q \phi: K \rightarrow Y$ is such that $(q \phi)^{-1} A=\phi^{-1}\left(q^{-1} A\right)$ is closed in $K$ since $A$ is compactly generated-the result follows.

Proposition 17.1.12. Let $X$ and $Y$ be $k$-spaces, and consider quotient maps $q: X \rightarrow X^{\prime}$ and $p: Y \rightarrow Y^{\prime}$. Then the product

$$
q \times p: X \times Y \rightarrow X^{\prime} \times Y^{\prime}
$$

is also a quotient map.
Proposition 17.1.13. A $k$-space $X$ is weak-Hausdorff if and only if the diagonal $\Delta X$ is closed in $X \times X$.

Proof. - $(\Leftarrow)$ Assume that $\Delta X$ is closed in $X \times X$. Let $K$ be a compact space and $\phi: K \rightarrow X$ a continuous map. Since $X$ is a $k$-space, we may simply show that $\phi K$ is compactly closed. To that end, let $C$ be another compact space together with a morphism $\psi: C \rightarrow X$, then

$$
\psi^{-1}(\phi K)=\pi_{2}(\phi \times \psi)^{-1}(\Delta X)
$$

is a closed subset of $C$-where $\pi_{2}: K \times C \rightarrow C$ is the canonical second projection. This proves that $\phi K$ is compactly closed.

- $(\Rightarrow)$ Suppose $X$ is a weak-Hausdorff $k$-space. It suffices to show that $\Delta X$ is compactly closed in $X \times_{c} X$, so that $\Delta X$ is closed in the $k$-space $X \times X$. Let $K$ be any compact set and $\phi: K \rightarrow X \times_{c} X$ be a morphism. Considering the canonical projections $\pi_{1}, \pi_{2}: X \times_{c} X \rightrightarrows X$, define the set

$$
A:=\pi_{1}(\phi K) \cup \pi_{2}(\phi K)
$$

of $X$. The set $A$ is constructed so that one has $\phi K \subseteq A \times{ }_{c} A$ and hence $\phi^{-1}(\Delta X)=$ $\phi^{-1}(\Delta A)$. Since $X$ is weak-Hausdorff, if we consider the maps $\pi_{1} \phi, \pi_{2} \phi: K \rightrightarrows X$, we find that $\pi_{1} \phi K$ and $\pi_{2} \phi K$ are both compact Hausdorff subspaces of $X$-hence $A$ is a compact Hausdorff space. Since $A$ is Hausdorff, it follows that $\Delta A$ is closed in $A \times_{c} A$, therefore by continuity $\phi^{-1}(\Delta A)=\phi^{-1}(\Delta X)$ is closed in $K$.

## CG Spaces

Definition 17.1.14 (Compactly generated space). A space $X$ is said to be compactly generated (or, shortly, CG space) if $X$ is a weak-Hausdorff $k$-space. The full-subcategory of Top composed of compactly generated spaces will be denoted by cgTop.

The $k$-ification functor $k$ :Top $\rightarrow k$-Top can also act on the category of weakHausdorff spaces wH -Top, producing compactly generated spaces:

$$
k: \mathrm{wH}-\mathrm{Top} \longrightarrow \mathrm{cgTop} .
$$

Lemma 17.1.15. The canonical forgetful functor $j$ : cgTop $\rightarrow$ wH-Top embedding CG spaces as weak-Hausdorff spaces is such that there exists a bijection

$$
\operatorname{Mor}_{c g T o p}(X, k Y) \simeq \operatorname{Mor}_{w H-T o p}(j X, Y)
$$

for every $X \in \mathrm{cgTop}$ and $Y \in \mathrm{wH}$-Top. Then $k$ is right-adjoint to $j$ :

$$
\text { cgTop } \underset{k}{\stackrel{j}{\rightleftarrows}} \text { wH-Top }
$$

Example 17.1.16. As examples of compactly generated spaces we have:
(a) If $X$ is locally compact, then it is a compactly generated space.
(b) If $X$ is a first-countable weak-Hausdorff space, then it is compactly generated.

Lemma 17.1.17. Let $X$ and $Y$ be topological spaces, then
(a) If $X$ is locally compact and $Y$ is compactly generated, then

$$
X \times Y=X \times_{c} Y
$$

(b) If both $X$ and $Y$ are weak-Hausdorff, then

$$
X \times Y=k X \times k Y
$$

(c) If both $X$ and $Y$ are compactly generated, then $X \times Y$ is a product in the category cgTop.

Lemma 17.1.18. Let $X$ be a compactly generated space. A set-function $f: X \rightarrow Y$ is continuous if and only if the restriction $\left.f\right|_{K}$ is continuous for every compact subspace $K \subseteq X$.

Proposition 17.1.19 (Quotients of CG spaces). Let $X$ be a compactly generated space, and $q: X \rightarrow Y$ be a quotient map. Then $Y$ is compactly generated if and only if the preimage of the diagonal of $Y$,

$$
(q \times q)^{-1}(\Delta Y)
$$

is closed in $X \times X$.
Proof. - $(\Rightarrow)$ Suppose $Y$ is compactly generated. From Proposition 17.1.13 we know that $\Delta Y$ is a closed subspace of $Y \times Y$, therefore by continuity of $q \times q$ with respect to the product topology $X \times_{c} X$ (see Proposition 13.3.7) we obtain that $(q \times q)^{-1}(\Delta Y)$ is a closed subspace of $X \times_{c} X$, it follows ${ }^{1}$ that $(q \times q)^{-1}(\Delta Y)$ is closed in $X \times X$.

- $(\Leftarrow)$ Suppose that $(q \times q)^{-1}(\Delta Y)$ is closed in $X \times X$. Since $q$ is a quotient map, then $Y$ is a quotient space of a $k$-space $X$-therefore $Y$ itself is a $k$-space.
Since $q$ is a quotient map and $X$ is in particular a $k$-space, from Proposition 17.1.12 we know that $q \times q: X \times X \rightarrow Y \times Y$ is a quotient map. Since $(q \times q)^{-1}(\Delta Y)$ is closed in $X \times X$ then $\Delta Y$ is closed in $Y \times Y$. Therefore by Proposition 17.1.13 we conclude that $Y$ is compactly generated.

[^24]Proposition 17.1.20. Let $X$ and $Y$ be compactly generated spaces, and $A \subseteq X$ be a closed subspace. Then for every morphism $f: A \rightarrow Y$ the attaching space $Y \cup_{f} X$ is compactly generated.

Definition 17.1.21 (Weak topology). Let $\left(X_{j}\right)_{j \in J}$ be a collection of topological spaces together with inclusions $X_{j} \hookrightarrow X_{j+1}$. We define the weak topology on the union set $X:=\bigcup_{j \in J} X_{j}$ as follows: a set $U \subseteq X$ is open if and only if $U \cap X_{j}$ is open in $X_{j}$ for all $j \in J$.

Proposition 17.1.22 (Colimits). Let $\left(X_{j}\right)_{j \in J}$ be a collection of compactly generated spaces together with inclusions $X_{j} \hookrightarrow X_{j+1}$ with closed images. Then the colimit

$$
\operatorname{colim}_{j \in J} X_{j}=\bigcup_{j \in J} X_{j}
$$

is compactly generated ${ }^{2}$.

### 17.2 Construction of CW-Complexes

Definition 17.2.1 (Attaching cells to a space). Let $f: \coprod_{j \in J} S_{j}^{n-1} \rightarrow X$ be a morphism of topological spaces, where $J$ is a set and $S_{j}^{n-1}$ is an indexed copy of the $(n-1)$-sphere. We shall consider the attaching space $Y$ given by the pushout

where $D_{j}^{n} \hookleftarrow S^{n-1}$ denotes an indexed copy of the $n$-disk. We define the following notions concerning the triple $(Y, X, f)$ :

- An $n$-cell in $Y$ is defined to be the image $e_{j}^{n}$ of a disk $D_{j}^{n}$ in $Y$, so that by construction we have $Y=X \cup\left(\bigcup_{j \in J} e_{j}^{n}\right)$.
- One can decompose $f$ into a collection of maps $\left(f_{j}^{n}: D_{j}^{n} \rightarrow Y\right)_{j \in J}$, each of these is called a characteristic map.

Definition 17.2.2 (Filtered space). A filtered topological space $X$ is a space together with an increasing sequence $\left(X_{n}\right)_{n \in \mathbf{N}}$ of closed subspaces, with inclusions $X_{n} \hookrightarrow X_{n+1}$ for all $n \in \mathbf{N}$, and

$$
X=\operatorname{colim}_{n \in \mathbf{N}} X_{n}=\bigcup_{n \in \mathbf{N}} X_{n}
$$

Where $X$ is endowed with the weak topology.

[^25]Definition 17.2.3 (Relative cell complex). A relative cell complex is a pair $(X, A)$ where $A$ is a $k$-space, and $X$ is a filtered space with:

- The initial space of the sequence $\left(X_{n}\right)_{n \in \mathbf{N}}$ associated to $X$ is

$$
X_{0}=A \amalg\left(\coprod_{j \in J_{0}} D_{j}^{0}\right),
$$

where $D_{j}^{0}$ is a copy of the 0 -disk indexed by a set $J_{0}$.

- For each $n \in \mathbf{N}$ we have an associated attaching map $f_{n}: \coprod_{j \in J_{n+1}} S_{j}^{n} \rightarrow X_{n}$ such that

$$
X_{n+1}=X_{n} \cup_{f_{n}}\left(\coprod_{j \in J_{n+1}} D_{j}^{n+1}\right)
$$

where $S_{j}^{n}$ and $D_{j}^{n+1}$ are copies of the $n$-sphere and $(n+1)$-disk indexed by a set $J_{n+1}$, respectively. In other words, the following diagram is a pushout in Top:


Each space $X_{n}$ is called the $n$-skeleton of $(X, A)$.
We say that $(X, A)$ is a finite relative cell complex if $X$ has finitely many $n$-cells for each $n \in \mathbf{N}$. Moreover, if $X=X_{n}$ we say that $(X, A)$ is an $n$-dimensional relative cell complex.

Together with relative cell complexes we define a cellular map $\phi:(X, A) \rightarrow(Y, B)$ between relative cell complexes to be a continuous map $\phi\left(X_{n}\right) \subseteq Y_{n}$ for each $n \in \mathbf{N}$. We say that a pair $(Y, A)$ is a subcomplex of $(X, A)$ if $Y$ is a subspace of $X$ that is given by the union of $A$ with cells of $X$.

A CW-complex is a relative cell complex $(X, \varnothing)$, that is, each skeleton $X_{n}$ is formed exclusively by attaching $n$-cells on $X_{n-1}$.

## Chapter 18

## Bundles

### 18.1 Bundle

## Initial Construction

Definition 18.1.1 (Bundle). A bundle is defined to be a triple ( $E, p, B$ ), where $E$ and $B$ are spaces and $p: E \rightarrow B$ is a morphism. We refer to $B$ as the base space, while $E$ is the total space, and $p$ is the projection of the bundle. As usual, given any $b \in B$, we name the object $p^{-1} b$ the fibre of the bundle over $b$.

A sub-bundle of $(E, p, B)$ is a bundle $\left(E^{\prime}, p^{\prime}, B^{\prime}\right)$ such that $E^{\prime}$ and $B^{\prime}$ are subspaces of $E$ and $B$, respectively, and $p^{\prime}=\left.p\right|_{E^{\prime}}: E^{\prime} \rightarrow B^{\prime}$.

Definition 18.1.2 (Product bundle). A product bundle over $B$ with fibre $F$ is a triple $(B \times F, p, B)$ where $p(x, y):=x$ is the first projection.

Definition 18.1.3 (Cross section). Given a bundle $(E, p, B)$, a cross section of the bundle is a section $s: B \rightarrow E$ of $p$-that is, $p s=\mathrm{id}_{B}$. As an immediate consequence of this definition, if $\left(E^{\prime}, p^{\prime}, B^{\prime}\right)$ is a sub-bundle of $(E, p, B)$, then $s$ is a cross section of $\left(E^{\prime}, p^{\prime}, B^{\prime}\right)$ if and only if $s B \subseteq E^{\prime}$.

Lemma 18.1.4 (Product bundle cross section). Given a product bundle ( $B \times F, p, B$ ), a cross section $s: B \rightarrow B \times F$ will always have the form $s=\operatorname{id}_{B} \times f$, where $f: B \rightarrow F$ is a uniquely defined morphism. Therefore the collection of cross sections of product bundles is in bijection with the collection of maps $B \rightarrow F$.

Proof. Let $s$ be any cross section of $(B \times F, p, B)$, from the definition of $s$, there exists unique morphisms $s^{\prime}: B \rightarrow B$ and $f: F \rightarrow B$ such that $s=s^{\prime} \times f$-it remains to be shown that $s^{\prime}$ is the identity on $B$. From definition of a product bundle, we know that $p s=s^{\prime}$-since $p$ is the projection of the first factor-moreover, from the definition of a cross section, $p s=\mathrm{id}_{B}$, therefore $s^{\prime}=\mathrm{id}_{B}$ as wanted.

Definition 18.1.5 (Stiefel variety). We define the Stiefel variety of orthonormal $k$-frames ${ }^{1}$ in $\mathbf{R}^{n}$ to be the compact subspace Stie $_{k} \mathbf{R}^{n} \subseteq\left(S^{n-1}\right)^{k}$ for which $\left(v_{1}, \ldots, v_{k}\right) \in \operatorname{Stie}_{k} \mathbf{R}^{n}$ if and only if $\left\langle v_{i}, v_{j}\right\rangle=\delta_{i j}$-where $\langle-,-\rangle$ is the standard euclidean inner product.

Definition 18.1.6 (Grassmann variety). The real $k$-Grassmann variety is defined to be the topological space $\operatorname{Grass}_{k} \mathbf{R}^{n}$ whose points are $k$-dimensional subspaces of $\mathbf{R}^{n}$, and endowed with the quotient topology generated by the map Stie ${ }_{k} \mathbf{R}^{n} \rightarrow \operatorname{Grass}_{k} \mathbf{R}^{n}$ given by $\left(v_{1}, \ldots, v_{k}\right) \mapsto\left\langle v_{1}, \ldots, v_{k}\right\rangle$. Since Stie ${ }_{k} \mathbf{R}^{n}$ is compact, it follows that Grass ${ }_{k} \mathbf{R}^{n}$ is also compact.

Example 18.1.7. Notice that Stie $_{1} \mathbf{R}^{n}=S^{n-1}$ and by the construction of the Grassmannian variety, we see that Grass $_{1} \mathbf{R}^{n}=\mathbf{R} \mathbf{P}^{n-1}$.

Definition 18.1.8 (Bundle morphism). Given two bundles ( $E, p, B$ ) and ( $E^{\prime}, p^{\prime}, B^{\prime}$ ), a morphism of bundles $(E, p, B) \rightarrow\left(E^{\prime}, p^{\prime}, B^{\prime}\right)$ is a pair $(u, f)$ of morphisms $u: E \rightarrow E^{\prime}$ and $f: B \rightarrow B^{\prime}$ such that the diagram

commutes in Top-equivalently, $u\left(p^{-1} B\right) \subseteq p^{-1}(f B)$. The special case where the base spaces coincide, we define a bundle morphism over $B$ (also referred to as $B$-morphism) $(E, p, B) \rightarrow\left(E^{\prime}, p^{\prime}, B\right)$ to be morphism $u: E \rightarrow E^{\prime}$ such that the triangle

commutes in Top, which can equivalently be expressed as the condition $u\left(p^{-1} B\right) \subseteq$ $p^{\prime-1} B$.

Given any two bundle morphisms $(u, f):(E, p, B) \rightarrow\left(E^{\prime}, p^{\prime}, B^{\prime}\right)$ and $(v, g):\left(E^{\prime}, p^{\prime}, B^{\prime}\right) \rightarrow$ $\left(E^{\prime \prime}, p^{\prime \prime}, B^{\prime \prime}\right)$, we define the composition of those morphisms to be the pair

$$
(v, g) \circ(u, f):=(v u, g f):(E, p, B) \longrightarrow\left(E^{\prime \prime}, p^{\prime \prime}, B^{\prime \prime}\right)
$$

which is again a bundle morphism, since

is a commutative diagram in Top.

[^26]Example 18.1.9 (Cross section as bundle morphism). Notice that a cross section is nothing more than a bundle morphism over $B$ of the form $s:\left(B, \mathrm{id}_{B}, B\right) \rightarrow(E, p, B)$.

Definition 18.1.10 (Category of bundles). We denote by Bun the category composed of bundles and bundle morphisms. Given a space $B$, we can also define a full subcategory $\mathrm{Bun}_{B}$ of Bun, whose objects are bundles with base space $B$ and bundle morphisms over B.

Definition 18.1.11 (Fibre of a bundle). We say that a space $F$ is the fibre of a bundle $(E, p, B)$ if there exists a topological isomorphism $p^{-1} b \simeq F$ for every $b \in B$. A bundle $(E, p, B)$ is said to be trivial with fibre $F$ if there exists a bundle $B$-isomorphism $(E, p, B) \simeq$ $(B \times F, p, B)$.

## Universal Properties

Proposition 18.1.12 (Products in Bun). Given a family of bundles $\left(E_{j}, p_{j}, B_{j}\right)_{j \in J}$, we define the product of this family of bundles to be the bundle

$$
\left(\prod_{j \in J} E_{j}, \prod_{j \in J} p_{j}, \prod_{j \in J} B_{j}\right),
$$

which is defines a product in the category Bun.
Proposition 18.1.13 (Pullbacks in Bun $\left._{B}\right)$. Given two bundles $\xi=(E, p, B)$ and $\xi^{\prime}=$ $\left(E^{\prime}, p^{\prime}, B\right)$, define

$$
E \oplus E^{\prime}:=\left\{\left(x, x^{\prime}\right) \in E \times E^{\prime}: p x=p^{\prime} x^{\prime}\right\}
$$

and $q: E \oplus E^{\prime} \rightarrow B$ to be the morphism $q\left(x, x^{\prime}\right):=p x=p^{\prime} x^{\prime}$. Then the triple

$$
\xi \oplus \xi^{\prime}:=\left(E \oplus E^{\prime}, q, B\right)
$$

called fibre product over $B$ of $\xi$ and $\xi^{\prime}$, is the pullback of the pair $\left(\xi, \xi^{\prime}\right)$ in the category $B u n_{B}$.

## Induced Bundle

Definition 18.1.14 (Induced bundle). Let $\xi=(E, p, B)$ be a bundle, and $f: B_{0} \rightarrow B$ be a continuous map. There exists an induced bundle of $\xi$ under $f$, denoted $f^{*} \xi$, with base space $B_{0}$ and total space $E_{0}$ defined as the pullback:


Explicitly, $E_{0}$ consists of pairs $\left(b_{0}, x\right) \in B_{0} \times E$ such that $f b_{0}=p x$. The projection of the bundle $f^{*} \xi$ is the map $p_{0}: E_{0} \rightarrow B_{0}$ given by $\left(b_{0}, x\right) \mapsto b_{0}$.

The mapping $f_{\xi}: E_{0} \rightarrow E$ given by $\left(b_{0}, x\right) \mapsto x$ induces a morphism of bundles $\left(f_{\xi}, f\right): f^{*} \xi \rightarrow \xi$, the so called canonical morphism of an induced bundle.

Proposition 18.1.15. Let $\xi=(E, p, B)$ be a bundle and $f: B_{0} \rightarrow B$ be a continuous map. Considering the canonical morphism $\left(f_{\xi}, f\right): f^{*} \xi \rightarrow \xi$, for each $b_{0} \in B_{0}$ the restricted map

$$
f_{\xi}: p_{0}^{-1} b_{0} \xrightarrow{\simeq} p^{-1}\left(f b_{0}\right)
$$

is a topological isomorphism. Furthermore, given a bundle $\eta=\left(E^{\prime}, p^{\prime}, B_{0}\right)$ and a mor$\operatorname{phism}(v, f): \eta \rightarrow \xi$, there exists a unique $B_{0}$-morphism $w: \eta \rightarrow f^{*} \xi$ such that $f_{\xi} w=v$. In other words, the following diagram

commutes in Top.
Proof. For the first part of the proposition, recalling the definition one has that the fibre $p_{0}^{-1} b_{0}$ is composed of pairs $\left(b_{0}, x\right) \in b_{0} \times E$ such that $p x=f b_{0}$-that is, $p_{0}^{-1} b_{0}=$ $b_{0} \times p^{-1}\left(f b_{0}\right)$. Since $f_{\xi}\left(b_{0}, x\right)=x$, then its restriction is a topological isomorphism with local inverse $x \mapsto\left(b_{0}, x\right)$.

To prove the second part, define a map $w: E^{\prime} \rightarrow E_{0}$ by $w:=\left(p^{\prime}, v\right)$, therefore

commutes, showing that $w$ is a $B_{0}$-morphism. Moreover, we have for any $y \in E^{\prime}$ that

$$
f_{\xi} w y=f_{\xi}\left(p^{\prime} y, v y\right)=v y
$$

therefore $f_{\xi} w=v$ as wanted. For the uniqueness of $w$, suppose $\ell: E^{\prime} \rightarrow E_{0}$ satisfies both $p_{0} \ell=p^{\prime}$ and $f_{\xi} \ell=v$, then since $\ell=\left(p_{0} \ell, f_{\xi} \ell\right)=\left(p^{\prime}, v\right)$, this shows that $\ell=w$. $\quad$.

Proposition 18.1.16. Given any continuous map $f: B_{0} \rightarrow B$, the induced bundle construction via $f$ is a functor

$$
f^{*}: \operatorname{Bun}_{B} \rightarrow \operatorname{Bun}_{B_{0}} .
$$

Moreover, given any morphism $u: \xi \rightarrow \eta$ in Bun $_{B}$, the diagram


Proof. For the functoriality, given any B-bundle morphism $u: \xi \rightarrow \eta$, the associated $\operatorname{map} f^{*} u: f^{*} \xi \rightarrow f^{*} \eta$ is canonically given by the mapping $\left(b_{0}, x\right) \mapsto\left(b_{0}, u x\right)$, which is a $B_{0}$-morphism of bundles. Moreover, if we consider the identity morphism id $\xi: \xi \rightarrow \xi$ one has

$$
f^{*}\left(\operatorname{id}_{\xi}\right)\left(b_{0}, x\right)=\left(b_{0}, x\right)=\operatorname{id}_{f^{*} \xi}\left(b_{0}, x\right),
$$

therefore $f^{*} \mathrm{id}_{\xi}=\mathrm{id}_{f^{*} \xi}$. Also, if $v: \eta \rightarrow \zeta$ is any other bundle morphism, we have

$$
f^{*}(v u)\left(b_{0}, x\right)=\left(b_{0}, v u x\right)=f^{*}(v)\left(b_{0}, u x\right)=f^{*}(v)\left(f^{*}(u)\left(b_{0}, x\right)\right),
$$

that is, $f^{*}(v u)=f^{*} v \circ f^{*} u$. This finishes the proof that $f^{*}$ is indeed a functor.
The only additional information the diagram brings is that $u f_{\xi}$ should equal $f_{\eta} f^{*} u$, and this is what we'll show. Let $\left(b_{0}, x\right) \in E\left(f^{*} \xi\right)$ be any point, then

$$
u f_{\xi}\left(b_{0}, x\right)=u x=f_{\eta}\left(b_{0}, u x\right)=f_{\eta}\left(f^{*} u\right)\left(b_{0}, x\right)
$$

which proves the commutativity of the diagram.
Proposition 18.1.17 (Functorial transitivity of the induced bundle). Consider continuous maps $B_{1} \xrightarrow{g} B_{0} \xrightarrow{f} B$, and a bundle $\xi=(E, p, B)$. Then the following are properties concerning the induced bundles over $\xi$ :
(a) Considering the identity map id: $B \rightarrow B$, there exists a bundle $B$-isomorphism

$$
\mathrm{id}^{*} \xi \simeq \xi
$$

(b) There exists a $B_{1}$-isomorphism of bundles

$$
g^{*} f^{*} \xi \simeq(f g)^{*} \xi
$$

Proof. Notice that the morphisms of bundles $\xi \rightarrow \mathrm{id}^{*} \xi$ given by $x \mapsto(p x, x)$ has an inverse $(b, x) \mapsto x$, proving the first isomorphism. For the second item, define $u: g^{*} f^{*} \xi \rightarrow(f g)^{*} \xi$ by the mapping $u\left(b_{1},\left(b_{0}, x\right)\right):=\left(b_{1}, x\right)$ then by the fact that $\left(b_{1},\left(b_{0}, x\right)\right) \in E\left(g^{*} f^{*} \xi\right)$ if and only if $g b_{1}=p_{1}\left(b_{0}, x\right)=b_{0}$, we can conclude that $u$ is an isomorphism of $B_{1}$-bundles.

### 18.2 Fibre Bundles

Definition 18.2.1 (Bundle projection). Let $X, B$, and $F$ be Hausdorff spaces. We say that a continuous map $p: X \rightarrow B$ is a bundle projection with fibre $F$ if for each $b \in B$ there exists a neighbourhood $U \subseteq B$ of $b$ such that there is a topological isomorphism

$$
\phi: U \times F \longrightarrow p^{-1} U, \quad \text { such that } \quad p \phi(x, y)=x
$$

for all $x \in U$ and $y \in F$-the map $\phi$ is called a trivialisation of the bundle over $U$. This means that on the set $p^{-1} U$, the map $p$ is a projection of the type $U \times F \rightarrow U$.

Definition 18.2.2 (Fibre bundle). Let $G$ be a topological group acting effectively on a Hausdorff space $F$-seen as a group of topological isomorphisms. Let $X$ and $B$ be Hausdorff spaces. We define a fibre bundle (or simply bundle) over the base space $B$ with total space $X$, fibre $F$, and structure group $G$, to be a pair $(p, \Phi)$ where $p: X \rightarrow B$ is a bundle projection and $\Phi$ is a collection of trivialisations of $p$ (as described in Definition 18.2.1)— the members of $\Phi$ will be called charts over $U$-such that:

- For each $b \in B$ there exists a neighbourhood $U \subseteq B$ of $b$ and a chart $\phi \in \Phi$ of the form $\phi: U \times F \rightarrow p^{-1} U$, where $\phi$ is a local trivialisation of $p$ over $U$.
- Given a chart $\phi: U \times F \rightarrow p^{-1} U$, member of $\Phi$, any subset $V \subseteq U$ is such that the restriction $\left.\phi\right|_{V \times F}$ belongs to the family $\Phi$.
- Given any pair of charts $\phi, \psi \in \Phi$ over a common open set $U$, there exists a continuous map $\theta: U \rightarrow G$ such that

$$
\psi(u, y)=\phi(u, \theta(u)(y)) .
$$

- The family $\Phi$ is maximal among the collections satisfying the previous properties. The fibre bundle is said to be smooth if each object above is a smooth manifold and all maps are smooth morphisms.


### 18.3 Vector Bundle

## First definitions

Definition 18.3.1 (Vector bundle). A (topological) vector bundle is a fibre bundle with a fibre $\mathbf{R}^{n}$ and structure group contained in $\mathrm{GL}_{n}(\mathbf{R})$. Given a vector bundle $\xi$, we denote its total space by $E \xi$ and base space by $B \xi$.

Notation 18.3.2. Given a vector bundle $\xi=(E, p, B)$, we denote by $\xi_{b}:=p^{-1} b$ (which can also be denoted by $E_{b}$ ) the fibre of $b \in B$ over $p$.

Definition 18.3.3 (Morphism of vector bundles). If $\xi=(E, p, B)$ and $\xi^{\prime}=\left(E^{\prime}, p^{\prime}, B^{\prime}\right)$ are any two vector bundles, we define a bundle morphism $\xi \rightarrow \xi^{\prime}$ is a pair $(u, f)$ of continuous maps $u: E \rightarrow E^{\prime}$ and $f: B \rightarrow B^{\prime}$ such that the diagram

commutes and the restriction $u_{b}: \xi_{b} \rightarrow \xi_{f b}^{\prime}$ is $\mathbf{R}$-linear for every $b \in B$.
Definition 18.3.4. We denote by VecBun the category of vector bundles and morphisms between them. Furthermore, given a base space $B$, we also define a full subcategory $\mathrm{VecBun}_{B}$ of vector bundles with base $B$ and $B$-morphisms.

## Charts and Atlases

Definition 18.3.5 (Vector bundle chart \& atlas). Given a vector bundle ( $E, p, B$ ), we define an $n$-dimensional vector bundle chart $(U, \phi)$, for some open set $U \subseteq B$, to be a topological isomorphism

$$
\phi: p^{-1} U \xrightarrow{\simeq} U \times \mathbf{R}^{n}
$$

such that the diagram

commutes in Top-and $\pi_{1}$ is the projection of the first factor. The isomorphism $\phi$ induces a collection $\left(\phi_{x}: p^{-1} x \xrightarrow{\simeq} \mathbf{R}^{n}\right)_{x \in U}$ of isomorphisms given by the composition


Therefore, given any $y \in p^{-1} U$, if $y \in p^{-1} x$, then $\phi y=\left(x, \phi_{x} y\right)$.
Definition 18.3.6. A family $\Phi:=\left(U_{j}, \phi_{j}\right)_{j \in J}$ of vector bundle charts on $(E, p, B)$ with domain covering $B$ and values in $\mathbf{R}^{n}$ is said to form a vector bundle atlas for $(E, p, B)$ if for any two vector bundle charts $(U, \phi)$ and $(V, \psi)$ of $\Phi$ we have:
(a) For every $x \in U \cap V$, the transition map

$$
\psi_{x} \phi_{x}^{-1}: \mathbf{R}^{n} \xrightarrow{\simeq} \mathbf{R}^{n}
$$

is an $\mathbf{R}$-linear topological isomorphism.
(b) The map $g: U \cap V \rightarrow \mathrm{GL}_{n}(\mathbf{R})$ sending $x \mapsto \psi_{x} \phi_{x}^{-1}$ is continuous.

If such conditions are satisfied, by the requirement of item (b), the atlas $\Phi$ induces a collection of continuous maps

$$
\left(g_{i j}: U_{i} \cap U_{j} \longrightarrow \mathrm{GL}_{n}(\mathbf{R})\right)_{(i, j) \in J \times J},
$$

called cocycle of $\Phi$. For any three $i, j, k \in J$, and $x \in U_{i} \cap U_{j} \cap U_{k}$, one has that

$$
g_{i j}(x) g_{j k}(x)=\left(\left(\phi_{i}\right)_{x}\left(\phi_{j}^{-1}\right)_{x}\right)\left(\left(\phi_{j}\right)_{x}\left(\phi_{k}^{-1}\right)_{x}\right)=\left(\phi_{i}\right)_{x}\left(\phi_{k}^{-1}\right)_{x}=g_{i k} x
$$

Moreover, for any $j \in J$ we find $g_{i i} x=\operatorname{id}_{\mathbf{R}^{n}}$. The tuple $(E, p, B, \Phi)$, is said to be a vector bundle with n-dimensional fibre. Furthermore an atlas for $(E, p, B)$ is a subatlas of $\Phi$.

For every $x \in B$, we can endow the fibre $E_{x}$ with the structure of an $\mathbf{R}$-vector space for which $\phi_{x}: E_{x} \xrightarrow{\simeq} \mathbf{R}^{n}$ is an $\mathbf{R}$-linear isomorphism, independently of the choice of $(\phi, U) \in \Phi$. We'll thus call $E$ an $n$-plane bundle

Definition 18.3 .7 (Zero section). Given a vector bundle $\xi$, we shall denote by Zero: $B \xi \rightarrow$ $E \xi$ the zero section of $\xi$, that is, the mapping $x \mapsto 0 \in E_{x}$.

Definition 18.3.8 (Trivial vector bundle). The $n$-dimensional trivial vector bundle is the vector bundle

$$
\varepsilon_{B}^{n}:=\left(B \times \mathbf{R}^{n}, p, B, \Phi\right),
$$

where $p$ is the projection of the first component, and $\Phi$ is the unique maximal vector bundle atlas on $\varepsilon_{B}^{n}$ containing identity maps for each open set of $B \times \mathbf{R}^{n}$. A vector bundle over $B$ is said to be trivial if it is isomorphic-such isomorphism is said to be a trivialisation-to $\varepsilon_{B}^{n}$ for some $n \in \mathbf{N}$.

Definition 18.3.9 (Smooth vector bundle). A vector bundle $\xi$ is said to be smooth if its associated spaces are smooth manifolds and the projection is a $C^{\infty}$-morphism.

Example 18.3.10 (Tangent bundle). Given a smooth $n$-manifold $M$, we define the tangent bundle (see the discussion at Section 23.4) of $M$ to be the vector bundle (TM, $\pi, M$ ). For each chart $\phi: U \rightarrow \mathbf{R}^{n}$ we define a vector bundle chart $\pi^{-1} U \rightarrow U \times \mathbf{R}^{n}$ mapping tangent vectors $X \mapsto\left(x, \phi_{* x} X\right)$.

Given any $C^{\infty}$-morphism $f: M \rightarrow N$ of manifolds, there is an induced vector bundle morphism $T f: T M \rightarrow T N$.

Theorem 18.3.11 (Isomorphism of vector bundles). Let $u: \xi \rightarrow \eta$ be a $B$-morphism of $n$-dimensional vector bundles. Then $u$ is a $B$-isomorphism of vector bundles if and only if $u_{b}: \xi_{b} \rightarrow \eta_{b}$ is an $\mathbf{R}$-linear isomorphism for each $b \in B$.

Proof. $(\Rightarrow)$ If $u$ is a $B$-isomorphism of vector bundles, then the restriction mapping $\left.u^{-1}\right|_{\eta_{b}}: \eta_{b} \rightarrow \xi_{b}$ is an inverse for $u_{b}$.
$(\Leftarrow)$ For the converse, suppose that the restriction $u_{b}: \xi_{b} \xrightarrow{\simeq} \eta_{b}$ is a linear isomorphism for all $b \in B$. Construct a set-function $v: E \eta \rightarrow E \xi$ such that $\left.v\right|_{\eta_{b}}:=u_{b}^{-1}$-we must show that $v$ is continuous. Let $U \subseteq B$ be any open set, and $\phi: \xi^{-1} U \xrightarrow{\leftrightharpoons} U \times \mathbf{R}^{n}$ and $\psi: \eta^{-1} U \xrightarrow{\simeq} U \times \mathbf{R}^{n}$ be vector bundle charts for $\xi$ and $\eta$, respectively. Then if we consider the map $\psi u \phi^{-1}$, we find that it has the form $(b, x) \mapsto\left(b, f_{b} x\right)$-where $b \mapsto f_{b}$ is a mapping $U \rightarrow \mathrm{GL}_{n}(\mathbf{R})$. On the other hand, the map $\phi v \psi^{-1}$ is of the form $(b, x) \mapsto\left(b, f_{b}^{-1} x\right)$-where $b \mapsto f$ is again a map $U \rightarrow \mathrm{GL}_{n}(\mathbf{R})$.

Definition 18.3.12 (Whitney sum). Given vector bundles $\xi, \eta \in \operatorname{VecBun}_{B}$, we define the Whitney sum of $\xi$ and $\eta$ to be the fibre product $\xi \oplus \eta \in \operatorname{VecBun}_{B}$, where $(\xi \oplus \eta)_{b}=\xi_{b} \oplus \eta_{b}$ has the structure of the direct sum of vector spaces. Given charts $\phi: U \times \mathbf{R}^{n} \rightarrow \xi^{-1} U$ and $\psi: U \times \mathbf{R}^{m} \rightarrow \eta^{-1} U$ for the respective vector bundles, we define an induced vector bundle chart for $\xi \oplus \eta$ to be

$$
\phi \oplus \psi: U \times \mathbf{R}^{n+m} \rightarrow \xi^{-1} U \oplus \eta^{-1} U
$$

## Induced Vector Bundles

Proposition 18.3.13 (Induced vector bundle). Let $\xi=(E, p, B)$ be an $n$-dimensional vector bundle, and $f: B_{0} \rightarrow B$ be a continuous map. Then the induced bundle $f^{*} \xi=$ $\left(E_{0}, p_{0}, B_{0}\right)$ admits a unique vector bundle structure, and the canonical morphism $\left(f_{\xi}, f\right): f^{*} \xi \rightarrow \xi$ is a vector bundle morphism. Moreover, $f_{\xi}: p_{0}^{-1} b_{0} \rightarrow \xi^{-1} b$ is a a linear isomorphism (refer to Definition 18.1.14).

Proof. From Definition 18.1.14 we know that $p_{0}^{-1} b_{0}=b_{0} \times p^{-1}\left(f b_{0}\right)$. Consider a pair of fibre points $\left(b_{0}, x\right),\left(b_{0}, y\right) \in p_{0}^{-1} b_{0}$ and define vector space structures $\left(b_{0}, x\right)+\left(b_{0}, y\right):=$ $\left(b_{0}, x+y\right)$, and $\lambda\left(b_{0}, x\right)=\left(b_{0}, \lambda x\right)$ for any $\lambda \in \mathbf{R}$. We know that the restriction $f_{\xi}: p_{0}^{-1} b_{0} \rightarrow p^{-1}\left(f b_{0}\right)$ is a topological isomorphism (see Proposition 18.1.15), but via the linear structure just defined in $p_{0}^{-1} b_{0}$, we see that $f_{\xi}$ is also a linear map, thus it's an isomorphism of vector spaces. This shows that $p_{0}^{-1} b_{0}$ is restrained to this linear structure, which proves uniqueness.

For a local trivialisation of $f^{*} \xi$, consider, for any open set $U \subseteq B$, a vector bundle chart $h: U \times \mathbf{R}^{n} \rightarrow p^{-1} U$ for $\xi$. We can construct a map $h^{\prime}: f^{-1} U \times \mathbf{R}^{n} \rightarrow p_{0}^{-1}\left(f^{-1} U\right)$ mapping $\left(b_{0}, x\right) \mapsto\left(b_{0}, h\left(f b_{0}, x\right)\right)$, which forms a vector bundle chart for $f^{*} \xi$ over the open set $f^{-1} U$.

## Homotopy Properties of Vector Bundles

Lemma 18.3.14 (Local trivialization). Let $\xi=(E, p, B \times I)$ be a $C^{r}$ vector bundle for some $0 \leqslant r \leqslant \infty$. Then every point $b \in B$ admits a neighbourhood $V \subseteq B$ such that $\left.\xi\right|_{V \times I}$ is trivial.

Proof. Since $\xi$ is locally trivial and $I$ is compact, then given $b \in B$ consider a neighbourhood $V_{j} \subseteq B$ of $b$ and a partition $0=t_{0}<\cdots<t_{m}=1$ of $I$ such that $\xi$ is trivial in a neighbourhood of $V_{j} \times I_{j}:=V_{j} \times\left[t_{j-1}, t_{j}\right]$ for each $0<j \leqslant m$. Define $V:=\bigcap_{j=1}^{m} V_{j}$ and let $\left(U_{j}\right)_{j=1}^{m}$ be a collection where $U_{j} \subseteq I$ is a neighbourhood of $I_{j}$ such that $\left.\xi\right|_{V \times U_{j}}$ is trivial.

We do induction on $m$. For the base case $m=1$ it follows by construction that $\left.\xi\right|_{V \times U_{1}}$ is trivial and since $I_{1}=I$ then $U_{1}=I$. Now if $m>1$ we can proceed assuming that the case is true for all $n<m$ : for each $1 \leqslant j \leqslant m$ we find a neighbourhood $J \subseteq I$ of the interval $I_{1} \cup \cdots \cup I_{j}=\left[0, t_{j}\right]$ for which $\left.\xi\right|_{V \times J}$ is trivial. This hints at the fact that it is sufficient to prove this construction for the case $m=2$, and this is what we'll set out to do.

Take two sub-intervals $U_{1}:=[0, b]$ and $U_{2}:=[a, 1]$ where $0<a<b<1$. Let $\phi_{j}: p^{-1}\left(V \times U_{j}\right) \rightarrow\left(V \times U_{j}\right) \times \mathbf{R}^{n}$, for $j \in\{1,2\}$, be $C^{r}$-charts (where we assumed $\xi$ to be $n$ dimensional). Associated with these maps is the cocycle mapping $g_{21}: V \times\left(U_{1} \cap U_{2}\right) \rightarrow$ $\mathrm{GL}_{n}(\mathbf{R})$ mapping $x \mapsto \phi_{1 x} \phi_{2 x}^{-1}$. Let $a<c<b$ and consider a $C^{r}$-map $\lambda: U_{2} \rightarrow U_{1} \cap U_{2}$ for which $\left.\lambda\right|_{[a, c]}=\operatorname{id}_{[a, c]}$. Defining $\mu:=\operatorname{id}_{V} \times \lambda: V \times[a, 1] \rightarrow V \times[a, b]$, we can construct a map

$$
h:=g \mu: V \times[a, 1] \longrightarrow \mathrm{GL}_{n}(\mathbf{R})
$$

such that $\left.h\right|_{V \times[a, c]}=\left.g\right|_{V \times[a, c]}$. This allows us to construct a trivialisation $\psi$ where we define, for each $x \in V \times I$, a map $\psi_{x}: \xi_{x} \rightarrow \mathbf{R}^{n}$, and

$$
\psi_{x}:= \begin{cases}\phi_{1 x}, & \text { if } x \in V \times[0, c] \\ h(x) \phi_{2 x}, & \text { if } x \in V \times[a, 1]\end{cases}
$$

so that by construction for any $x \in[a, c]$ one has $h(x) \phi_{2 x}=\left(\phi_{1 x} \phi_{2 x}^{-1}\right) \phi_{2 x}=\phi_{1 x}$, showing that $\psi_{x}$ is well defined. This defines a trivialisation $\psi: p^{-1}(V \times I) \rightarrow(V \times I) \times \mathbf{R}^{n}$ as wanted.

Corollary 18.3.15. Any $C^{r}$ vector bundle, with $0 \leqslant r \leqslant \infty$, over an interval is trivial

## Oriented Vector Bundles

Definition 18.3.16. Let $\xi$ be a vector bundle. We define an orientation for $\xi$ to be a family $\omega=\left(\omega_{x}\right)_{x \in B \xi}$ where $\omega_{x}$ is an orientation for the vector space fibre $\xi_{x}$ such that $\xi$ has an atlas for which: every chart $\phi:\left.\xi\right|_{U} \rightarrow \mathbf{R}^{n}$ in the atlas $\Phi$ of $\xi$ has an orientation preserving map $\phi_{x}:\left(\xi_{x}, \omega_{x}\right) \rightarrow\left(\mathbf{R}^{n}, \omega_{n}\right)^{2}$. If this is the case, $\omega$ is said to be a coherent family of orientations of the fibres and $\Phi$ is an oriented atlas of $\xi$.

Lemma 18.3.17. Let $f: \eta \xrightarrow{\simeq} \xi$ be an isomorphism in VecBun, and let $\omega$ be an orientation for $\xi$. Then there exists a unique orientation $\theta$ for $\eta$ for which $f$ preserves the orientations of the fibres.

Proposition 18.3.18. Every vector bundle over a simply connected manifold admits an orientation.

Proposition 18.3.19. A vector bundle $\xi$ over a manifold $M$ is orientable if and only if every loop $\gamma \in \Omega M$ preserves the orientation of $\xi_{\gamma(0)}$.

Corollary 18.3.20. An orientable vector bundle over a connected manifold has only two orientations.

[^27]
## Chapter 19

## Homology

### 19.1 Singular Homology

## Life in Topological Spaces: Singular Simplices

Definition 19.1.1 (Convexity). The following are definitions concerning convexity of spaces in $\mathbf{R}^{n}$ :
(a) Given any two points $x, y \in \mathbf{R}^{n}$, we define the segment from $x$ to $y$ to be the set of points $\{(1-t) x+t y: t \in[0,1]\}$.
(b) A set $C \subseteq \mathbf{R}^{n}$ is said to be convex if for any pair of points $x, y \in C$ the segment from $x$ to $y$ is contained in $C$.
(c) Given a set $A \subseteq \mathbf{R}^{n}$, we define the convex hull of $A$ to be the intersection of all convex sets of $\mathbf{R}^{n}$ containing $A$.
(d) An $m$-simplex $\sigma$ in $\mathbf{R}^{n}$ is defined to be the convex-hull of a collection of $m+1$ distinct points $\left\{x_{0}, \ldots, x_{m}\right\}$ such that the set $\left\{x_{j}-x_{0}: 1 \leqslant j \leqslant m\right\}$ is linearly independent. The points $x_{j}$ are called the vertices of $\sigma$. If an order is assigned to the collection of vertices of $\sigma$, then we obtain an ordered simplex-the ordered simplex is commonly written as $\sigma=\left[x_{0}, \ldots, x_{m}\right]$.

Lemma 19.1.2. Let $\left\{x_{0}, \ldots, x_{m}\right\} \subseteq \mathbf{R}^{n}$ be a collection of $m+1$ distinct points. Then the set $L=\left\{x_{j}-x_{0}: 1 \leqslant j \leqslant m\right\}$ is linearly independent if and only if for any two sequences of parameters $\left(s_{j}\right)_{j=0}^{m}$ and $\left(t_{j}\right)_{j=0}^{m}$ satisfying both $\sum_{j} s_{j} x_{j}=\sum_{j} t_{j} x_{j}$ and $\sum_{j} s_{j}=\sum_{j} t_{j}$ implies that $s_{j}=t_{j}$ for each $0 \leqslant j \leqslant m$.
pair of sequences of parameters. Then

$$
\begin{aligned}
0 & =\sum_{j=0}^{m}\left(s_{j}-t_{j}\right) x_{j} \\
& =\sum_{j=0}^{m}\left(s_{j}-t_{j}\right) x_{j}-\left(\sum_{j=0}^{m} s_{j}-t_{j}\right) x_{0} \\
& =\sum_{j=1}^{m}\left(s_{j}-t_{j}\right)\left(x_{j}-x_{0}\right)
\end{aligned}
$$

but since $L$ is linearly independent, then it must be the case that $s_{j}-t_{j}=0$. For the second case, assume only that the two conditions for any pair sequences are met: then if $\sum_{j=1}^{m} a_{j}\left(x_{j}-x_{0}\right)=0$ then $\sum_{j} a_{j} x_{j}=\sum_{j} a_{j} x_{0}$ hence $a_{j}=0$ for each $1 \leqslant j \leqslant m$-proving that $L$ is linearly independent.

Corollary 19.1.3. Let $\sigma$ be the $m$-simplex given by the convex hull of the set of points $\left\{x_{0}, \ldots, x_{m}\right\}$. Then every point contained in $\sigma$ is uniquely represented by $\sum_{j=0}^{m} t_{j} x_{j}$ where $t_{j} \geqslant 0$ and $\sum_{j} t_{j}=1$. The tuple $\left(t_{0}, \ldots, t_{m}\right)$ is called the barycentric coordinates of the point in question.

Corollary 19.1.4. Let $\sigma=\left[x_{0}, \ldots, x_{m}\right]$ be an ordered $m$-simplex. Then the map $\Delta_{\text {top }}^{m} \rightarrow$ $\sigma$, from the standard topological $m$-simplex, given by $\left(t_{0}, \ldots, t_{m}\right) \mapsto \sum_{j=0}^{m} t_{j} x_{j}$ is an isomorphism of topological spaces.

Definition 19.1.5 (Singular simplex). Let $X$ be a topological space. We define a singular $m$-simplex in $X$ to be a morphism of topological spaces $\phi: \Delta_{\text {top }}^{m} \rightarrow X$.

Following the construction of the singular complex functor (see Definition 20.2.19) we define a structure of a free abelian group to the set

$$
\operatorname{Sing}_{m} X:=\operatorname{Sing}_{\Delta_{\text {top }}^{m}} X=\operatorname{Mor}_{\text {Top }}\left(\Delta_{\text {top }}^{m}, X\right)
$$

An element of $\operatorname{Sing}_{m} X$ is called a singular m-chain of $X$ and assumes the form $\sum_{\phi} n_{\phi} \phi$ for finitely many non-zero integers $n_{\phi}$ associated with singular $m$-simplices $\phi$.

Together with the singular simplex we also define, for each $0 \leqslant j \leqslant m$ a morphism of abelian groups

$$
\partial_{m}^{(j)}: \operatorname{Sing}_{m} X \longrightarrow \operatorname{Sing}_{m-1} X
$$

called $j$-th face map, which is explicitly given by

$$
\partial_{m}^{(j)} \phi\left(t_{0}, \ldots, t_{m-1}\right):=\phi\left(t_{0}, \ldots, t_{j-1}, 0, t_{j}, \ldots, t_{m-1}\right) .
$$

That is, the $j$-th face map embeds $\Delta_{\text {top }}^{m-1}$ into $\Delta_{\text {top }}^{m}$ face opposite to the $j$-th vertex and then map it to $X$ again via $\phi$. More compactly, we can also write

$$
\partial_{m}^{(j)} \phi=:\left.\phi\right|_{[0, \ldots, \widehat{j}, \ldots, m]} .
$$

We shall also define the boundary operator to be the morphism of abelian groups

$$
\partial_{m}: \operatorname{Sing}_{m} X \longrightarrow \operatorname{Sing}_{m-1} X
$$

to be given by the alternating sum of face maps

$$
\partial_{m}:=\sum_{j=0}^{m}(-1)^{j} \partial_{m}^{(j)}
$$

Proposition 19.1.6. The composition of the maps $\operatorname{Sing}_{m} X \xrightarrow{\partial_{m}} \operatorname{Sing}_{m-1} X \xrightarrow{\partial_{m-1}} \operatorname{Sing}_{m-2} X$ is identically zero. That is, the boundary of an $m$-chain is an $(m-1)$-chain with empty boundary.

Proof. With no loss of generality, consider merely one singular $m$-simplex $\phi$. One has

$$
\begin{aligned}
\partial_{m-1} \partial_{m} \phi & =\sum_{i=0}^{m}(-1)^{i}\left(\sum_{j=0}^{m-1}(-1)^{j} \partial_{j} \partial_{i} \phi\right) \\
& =\sum_{i=0}^{m}(-1)^{i}\left(\left.\sum_{i>j}(-1)^{j} \phi\right|_{\left[v_{0}, \ldots, \widehat{v}_{j}, \ldots, \widehat{v}_{i}, \ldots, v_{m-2}\right]}+\left.\sum_{i<j}(-1)^{j-1} \phi\right|_{\left[v_{0}, \ldots, \widehat{v}_{i}, \ldots, \widehat{v}_{j}, \ldots, v_{m-2}\right]}\right)
\end{aligned}
$$

Notice that the first term accounts for the removal of $j$ then of $i$, with $i>j$, yielding a factor $(-1)^{i+j}$ for each newly generated simplex. The second term deals with the removal of $i$ first and then $j$, with $i<j$, thus $j$ must be updated to a lower index by 1 , hence the factor $(-1)^{i+j-1}$. Notice that the pair of sums has different signs and the same simplices-therefore they pairwise cancel, yielding zero.

## Cycles \& Boundaries: Homology

Definition 19.1.7. Let $X$ be a topological space. We define the following:
(a) A singular $n$-chain $c \in \operatorname{Sing}_{n} X$ is said to be an $n$-cycle if $\partial c=0$. The collection of $n$-cycles of $\operatorname{Sing}_{n} X$ forms a subgroup $Z_{n}(X)$.
(b) A singular $n$-chain $b \in \operatorname{Sing}_{n} X$ is called an $n$-boundary if there exists $s \in \operatorname{Sing}_{n+1} X$ with $\partial s=b$. The collection of such $n$-boundaries forms a subgroup of $\operatorname{Sing}_{n} X$ named $\mathrm{B}_{n}(X)$
(c) From Proposition 19.1.6 we obtain $B_{n}(X) \subseteq Z_{n}(X)$ and thus we may define the $n$-th singular homology group of $X$ as the quotient group

$$
\mathrm{H}_{n}(X):=\mathrm{Z}_{n}(X) / \mathrm{B}_{n}(X) .
$$

Definition 19.1.8 (Chain complex). We define a chain complex to be a Z-graded abelian group $C=\left(C_{n}\right)_{n \in \mathbf{Z}}$ together with a graded endomorphism $\partial: C \rightarrow C$ of degree -1 such
that $\partial \circ \partial=0$. Given chain complexes $(C, \partial)$ and $\left(C^{\prime}, \partial^{\prime}\right)$ we define a morphism of chain complexes (or chain map) $\Phi: C \rightarrow C^{\prime}$ to be a morphism of graded abelian groups with zero degree satisfying the commutativity of the square

for every $n \in \mathbf{Z}$.
Considering the Z-graded abelian groups $Z_{\bullet}(C)=\operatorname{ker} \partial$ and $B_{\bullet}(C)=\operatorname{im} \partial$ we obtain the following Z-graded quotient abelian group

$$
\mathrm{H}_{\bullet}(\mathrm{C})=\mathrm{Z}_{\bullet}(\mathrm{C}) / \mathrm{B}_{\bullet}(\mathrm{C}) .
$$

Moreover, given a chain map $\Phi: C \rightarrow C^{\prime}$ since $\Phi\left(Z_{\bullet}(C)\right) \subseteq Z_{\bullet}\left(C^{\prime}\right)$ and $\Phi\left(B_{\bullet}(C)\right) \subseteq B_{\bullet}(C)$ then $\Phi$ induces a morphism of graded groups between the homology groups associated with $C$ and $C^{\prime}$ :

$$
\Phi_{*}: \mathrm{H}_{\bullet}(C) \longrightarrow \mathrm{H}_{\bullet}\left(C^{\prime}\right)
$$

Given a topological space $X$, its associated chain complex is given by Sing. $X$ together with the boundary operator $\partial$ and the homology of $X$ is the homology of the chain complex (Sing. $X, \partial$ ).

Corollary 19.1.9. Given a topological space $X$, the homology $H_{\bullet}(X)$ is zero if and only if the sequence (Sing. $X, \partial_{\bullet}$ ) is exact. In other words, $H_{\bullet}(X)$ measures how exact is the chain complex of $X$.

Proposition 19.1.10. Let $f: X \rightarrow Y$ be a morphism of topological spaces. For each $n \in \mathbf{Z}$ and singular $n$-simplex $\phi \in \operatorname{Sing}_{n} X$ we obtain an induced singular $n$-simplex in $Y$ given by the pushforward $f_{\# n}(\phi)=f \phi \in \operatorname{Sing}_{n} Y$. From this we obtain a morphism of abelian groups $f_{\# n}: \operatorname{Sing}_{n} X \rightarrow \operatorname{Sing}_{n} Y$. This can be extended to a chain map

$$
f_{\#}: \text { Sing. } X \longrightarrow \text { Sing. }_{\bullet} Y
$$

Moreover, this map also induces a graded morphism of zero degree between the homology groups of $X$ and $Y$ :

$$
f_{*}: \mathrm{H}_{\bullet}(X) \longrightarrow \mathrm{H}_{\bullet}(Y)
$$

Both $f_{\#}$ and $f_{*}$ have a functorial nature.
Proof. To prove that $f_{\#}$ is a chain map, it must be the case that

commutes for any $n \in \mathbf{Z}$. Let $\phi \in \operatorname{Sing}_{n} X$ be any element and notice that for any $0 \leqslant j \leqslant n$ we have

$$
f_{\# n-1}\left(\partial_{n}^{(j)} \phi\right)=\left.f \phi\right|_{\left[0, \ldots, \widehat{j}_{, \ldots, n}\right]^{\prime}}
$$

moreover

$$
\begin{equation*}
\partial_{n}^{\prime}(j) f_{\# n}(\phi)=\left.\left(f_{\# n}(\phi)\right)\right|_{[0, \ldots, \widehat{j}, \ldots, n]}=\left.(f \phi)\right|_{[0, \ldots, \widehat{j}, \ldots, n]}=\left.f \phi\right|_{[0, \ldots, \widehat{j}, \ldots, n]} \tag{দ}
\end{equation*}
$$

therefore the diagram indeed commutes.
Proposition 19.1.11. Let $\left(C^{\lambda}, \partial^{\lambda}\right)_{\lambda \in \Lambda}$ be a collection of chain complexes and define a chain complex $C:=\bigoplus_{\lambda \in \Lambda} C^{\lambda}$ as $C_{p}=\bigoplus_{\lambda \in \Lambda} C_{p}^{\lambda}$ and the natural boundary maps. Then there exists a natural isomorphism

$$
\mathrm{H}_{\bullet}(C) \simeq \bigoplus_{\lambda \in \Lambda} \mathrm{H}_{\bullet}\left(C^{\lambda}\right)
$$

Proof. From construction we have Sing. $\mathbf{C}=\bigoplus_{\lambda \in \Lambda}$ Sing. $C^{\lambda}$. Therefore for any $p \in \mathbf{Z}$ we have

$$
\begin{aligned}
\mathrm{H}_{p}(C) & =\mathrm{Z}_{p}(C) / \mathrm{B}_{p}(C) \\
& =\frac{\bigoplus_{\lambda \in \Lambda} \mathrm{Z}_{p}\left(C^{\lambda}\right)}{\bigoplus_{\lambda \in \Lambda} \mathrm{B}_{p}\left(C^{\lambda}\right)} \\
& \simeq \bigoplus_{\lambda \in \Lambda} \mathrm{Z}_{p}\left(C^{\lambda}\right) / \mathrm{B}_{p}\left(C^{\lambda}\right) \\
& =\bigoplus_{\lambda \in \Lambda} \mathrm{H}_{p}\left(C^{\lambda}\right)
\end{aligned}
$$

Proposition 19.1.12. Let $X \simeq \coprod_{\lambda \in \Lambda} X_{\lambda}$ be a topological space, where $X_{\lambda}$ are pathconnected components of $X$, then

$$
\mathrm{H}_{\bullet}(X) \simeq \bigoplus_{\lambda \in \Lambda} \mathrm{H}_{\bullet}\left(X_{\lambda}\right)
$$

Proof. For each $p \in \mathbf{Z}$ consider the morphism of groups $\Theta_{p}: \bigoplus_{\lambda} \operatorname{Sing}_{p} X_{\lambda} \rightarrow \operatorname{Sing}_{p} X$ given by

$$
\left(\sum_{\phi_{\lambda}} n_{\lambda, \phi_{\lambda}} \phi\right)_{\lambda} \longmapsto \sum_{\lambda} \sum_{\phi_{\lambda}} n_{\lambda, \phi_{\lambda}} \phi_{\lambda} .
$$

The Z-graded abelian group morphism $\Theta: \bigoplus_{\lambda}$ Sing. $X_{\lambda} \rightarrow$ Sing. $X$ given by $\left(\Theta_{p}\right)_{p \in \mathbf{Z}}$ is a morphism between free abelian groups, therefore $\Theta$ is injective. Let $\phi \in \operatorname{Sing}_{p} X$ be any singular $p$-simplex of $X$. Since $\Delta_{\text {top }}^{p}$ is path-connected, its image $\phi\left(\Delta_{\text {top }}^{p}\right) \subseteq X$ must be contained in some unique path-connected component $X_{\lambda}$ of $X$. Therefore we might as well restrict the target of $\phi$ and obtain a singular $p$-simplex $\left.\phi\right|^{X_{\lambda}} \in \operatorname{Sing}_{p} X_{\lambda}$. This shows that $\Theta_{p}\left(\left.\phi\right|^{X_{\lambda}}\right)=\phi$ and therefore $\Theta$ is a surjective map. Therefore $\Theta$ is an isomorphism. Using Proposition 19.1.11 we obtain the desired isomorphism between homology groups.

### 19.2 Calculating The Homology: Some Examples

Proposition 19.2.1. Let $X$ be path-connected, then $H_{0}(X) \simeq \mathbf{Z}$.
Proof. Define the morphism of groups $\varepsilon$ : $\operatorname{Sing}_{0} X \rightarrow \mathbf{Z}$ to be given by $\sum_{j} n_{j} \sigma_{j} \mapsto \sum_{j} n_{j}-$ which is clearly surjective. Notice that if $\sigma \in \operatorname{Sing}_{1} X$ is any singular 1-simplex, then $\partial_{1} \sigma=\sigma_{[0,1]}-\sigma_{[0, \widehat{1}]}$. Therefore $\varepsilon \sigma=1-1=0$ and thus $\operatorname{im} \partial_{1} \subseteq \operatorname{ker} \varepsilon$. For the contrary, let $\sum_{j} n_{j} \sigma_{j} \in \operatorname{ker} \varepsilon$-that is, $\sum_{j} n_{j}=0$. Let $x_{0} \in X$ be any point and, since $X$ is pathconnected, take for each $j$ a singular 1-simplex $\tau_{j}: \Delta_{\text {top }}^{1} \rightarrow X$ (which is a path in $X$ since $\left.\Delta_{\text {top }}^{1}=I\right)$ connecting $x_{0}$ with im $\sigma_{j}$-which is a unique point in $X$ the domain of $\sigma_{j}$ is a single point $\Delta_{\text {top }}^{0}$. Then

$$
\partial_{1}\left(\sum_{j} n_{j} \tau_{j}\right)=\sum_{j} n_{j}\left(\sigma_{j}-x_{0}\right)=\sum_{j} n_{j} \sigma_{j}-\left(\sum_{j} n_{j}\right) x_{0}=\sum_{j} n_{j} \sigma_{j}
$$

which shows that $\operatorname{ker} \varepsilon \subseteq \operatorname{im} \partial_{1}$. Therefore

induces a unique isomorphism $H_{0}(X) \simeq \mathbf{Z}$ since

$$
\operatorname{Sing}_{0}(X) / \operatorname{ker} \varepsilon=\operatorname{Sing}_{0}(X) / \operatorname{im} \partial_{1}=\mathrm{H}_{0}(X)
$$

Example 19.2.2 (Single point space homology). Let $X=\{*\}$ be the discrete space with a single point. For any $p \in \mathbf{N}$ we have a unique singular $p$-simplex $\phi_{p}: \Delta_{\text {top }}^{p} \rightarrow X$, and thus $\partial_{p}^{(j)} \phi_{p}=\phi_{p-1}$-where $\phi_{p-1}$ is the unique singular $p-1$-simplex of $X$. This shows that each $\operatorname{Sing}_{p} X$ is a cyclic group generated by $\phi_{p}$. Moreover, for any $p \in \mathbf{N}$ we have

$$
\partial \phi_{p}=\sum_{j=0}^{p}(-1)^{j} \partial_{j} \phi_{p}=\sum_{j=0}^{p}(-1)^{j} \phi_{p-1}= \begin{cases}\phi_{p-1}, & \text { if } p \text { is even and } p>0 \\ 0, & \text { otherwise }\end{cases}
$$

Therefore one has $Z_{p}(X)=B_{p}(X)$ for each $p>0$, and $Z_{0}(X)=\operatorname{Sing}_{0} X$ while $B_{0}(X)=0$. Hence we conclude that the homology groups associated to $X$ are

$$
\mathrm{H}_{p}(X)= \begin{cases}\mathbf{Z}, & \text { if } p=0 \\ 0, & \text { if } p>0\end{cases}
$$

where $H_{0}(X)=\mathbf{Z}$ comes from the fact that $X$ is path-connected and Proposition 19.2.1.
Theorem 19.2.3. Let $X \subseteq \mathbf{R}^{n}$ be a convex subspace. Then for any $p>0$ we have

$$
\mathrm{H}_{p}(X)=0 .
$$

Proof. If $X=\varnothing$ then we are done, now assume that $X$ is non-empty. Let $x \in X$ be any point and consider a singular $p$-simplex $\phi \in \operatorname{Sing}_{p} X$ for some $p \geqslant 0$. Since $X$ is convex, we can define a singular $(p+1)$-simplex $\theta \in \operatorname{Sing}_{p+1} X$ given by

$$
\theta\left(t_{0}, \ldots, t_{p+1}\right):= \begin{cases}\left(1-t_{0}\right)\left(\phi\left(\frac{t_{1}}{1-t_{0}}, \ldots, \frac{t_{p+1}}{1-t_{0}}\right)\right)+t_{0} x, & \text { for } t_{0}<1 \\ x, & \text { for } t_{0}=1\end{cases}
$$

We have to check that $\theta$ is indeed continuous at $x$. Notice that

$$
\begin{aligned}
\lim _{t_{0} \rightarrow 1}\left\|\theta\left(t_{0}, \ldots, t_{p+1}\right)-x\right\| & =\lim _{t_{0} \rightarrow 1}\left\|\left(1-t_{0}\right)\left(\phi\left(\frac{t_{1}}{1-t_{0}}, \ldots, \frac{t_{p+1}}{1-t_{0}}\right)\right)-\left(1-t_{0}\right) x\right\| \\
& \leqslant \lim _{t_{0} \rightarrow 1}\left(1-t_{0}\right)\left(\left\|\phi\left(\frac{t_{1}}{1-t_{0}}, \ldots, \frac{t_{p+1}}{1-t_{0}}\right)\right\|+\|x\|\right) .
\end{aligned}
$$

Using the fact that the image $\phi\left(\Delta_{\text {top }}^{p}\right) \subseteq X$ is compact, it follows that the second term of the right-hand side of the above equation is bound, therefore

$$
\lim _{t_{0} \rightarrow 1}\left\|\theta\left(t_{0}, \ldots, t_{p+1}\right)-x\right\|=0
$$

This construction induces a morphism $T_{p}: \operatorname{Sing}_{p} X \rightarrow \operatorname{Sing}_{p+1} X$ given by $\phi \mapsto \theta$ for every $p \geqslant 0$. Moreover, notice that $\partial_{p+1}^{0} T_{p} \phi=\partial_{p+1}^{0} \theta=\phi$ and $T_{p} \partial_{p+1}^{0} \theta=T_{p} \phi=\theta$, therefore the induced morphism $T$ : Sing. $X \rightarrow$ Sing. $X$ is an isomorphism of Z-graded abelian groups with inverse $\partial_{0}^{0}$. Hence $\operatorname{Sing}_{p} X \simeq \operatorname{Sing}_{p+1} X$ for any $p \geqslant 0$.

We now check that $T$ is a chain map. For each $1 \leqslant j \leqslant p+1$ one has

$$
\begin{aligned}
\partial_{p+1}^{j}\left(T_{p}(\phi)\right)\left(t_{0}, \ldots, t_{p}\right) & =T_{p}(\phi)\left(t_{0}, \ldots, t_{j-1}, 0, t_{j}, \ldots, t_{p}\right) \\
& =\left(1-t_{0}\right)\left(\phi\left(\frac{t_{1}}{1-t_{0}}, \ldots, \frac{t_{j-1}}{1-t_{0}}, 0, \frac{t_{j}}{1-t_{0}}, \ldots, \frac{t_{p}}{1-t_{p}}\right)\right)+t_{0} x
\end{aligned}
$$

and for $\partial_{p}^{j-1}$ we have

$$
\begin{aligned}
T_{p-1}\left(\partial_{p}^{j-1} \phi\right)\left(t_{0}, \ldots, t_{p}\right) & =\left(1-t_{0}\right) \partial_{p}^{j-1} \phi\left(\frac{t_{1}}{1-t_{0}}, \ldots, \frac{t_{p}}{1-t_{0}}\right)+t_{0} x \\
& =\left(1-t_{0}\right) \phi\left(\frac{t_{1}}{1-t_{0}}, \ldots, \frac{t_{j-1}}{1-t_{0}}, 0, \frac{t_{j}}{1-t_{0}}, \ldots, \frac{t_{p}}{1-t_{0}}\right)+t_{0} x .
\end{aligned}
$$

With this we conclude that

$$
\partial_{p+1}^{j} T_{p}=T_{p-1} \partial_{p}^{j-1}
$$

For any $\phi \in \operatorname{Sing}_{p} X$ we have

$$
\begin{aligned}
\partial_{p+1} T_{p} \phi & =\partial_{p+1}^{0} T_{p} \phi+\sum_{j=1}^{p+1}(-1)^{j} \partial_{p+1}^{j} T_{p} \phi \\
& =\partial_{p+1}^{0} T_{p} \phi+\sum_{j=1}^{p+1}(-1)^{j} \partial_{p+1}^{j} T_{p} \phi-\left(\sum_{j=1}^{p+1}(-1)^{j} T_{p-1} \partial_{p}^{j-1} \phi+\sum_{j=0}^{p}(-1)^{j} T_{p-1} \partial_{p}^{j} \phi\right) \\
& =\partial_{p+1}^{0} T_{p} \phi+\left(\sum_{j=1}^{p+1}(-1)^{j} \partial_{p+1}^{j} T_{p} \phi-\sum_{j=1}^{p+1}(-1)^{j} T_{p-1} \partial_{p}^{j-1} \phi\right)+\sum_{j=0}^{p}(-1)^{j} T_{p-1} \partial_{p}^{j} \phi \\
& =\partial_{p+1}^{0} T_{p} \phi+\sum_{j=0}^{p}(-1)^{j} T_{p-1} \partial_{p}^{j} \phi \\
& =\phi-T \partial_{p} \phi
\end{aligned}
$$

Therefore we conclude that for any $p \geqslant 1$ we have

$$
\partial_{p+1} T_{p}+T_{p-1} \partial_{p}=\operatorname{id}_{\text {Sing }_{p} X}
$$

Consider a cycle $z \in Z_{p}(X)$ for $p>0$ and notice that $\partial_{p+1} T_{p} z+T_{p-1} \partial_{p} z=z$ implies in $\partial_{p+1} T_{p} z=z$ since $\partial_{p} z=0$. Therefore from definition we conclude that $z \in \mathrm{~B}_{p}(X)$, which implies that $H_{p}(X)=0$ for each $p>0$.

### 19.3 Chain Homotopies

Let $T:(C, \partial) \rightarrow\left(C^{\prime}, \partial^{\prime}\right)$ be a morphism of graded groups of degree 1 . The morphism of graded groups

$$
\partial^{\prime} T+T \partial: C \longrightarrow C^{\prime}
$$

is then of degree 0 since both $\partial$ and $\partial^{\prime}$ have degree -1 . Moreover, $\partial^{\prime} T+T \partial$ is in fact a chain map since

$$
\partial^{\prime}\left(\partial^{\prime} T+T \partial\right)=\partial^{\prime} \partial^{\prime} T+\partial^{\prime} T \partial=\partial^{\prime} T \partial=\partial^{\prime} T \partial+T \partial \partial=\left(\partial^{\prime} T+T \partial\right) \partial
$$

shows that it satisfies the needed commutativity condition. Notice that for any cycle $z \in Z_{p}(C)$ we have

$$
\left(\partial^{\prime} T+T \partial\right) z=\partial^{\prime} T z+T \partial z=\partial^{\prime} T z \in \mathrm{~B}_{p}\left(C^{\prime}\right)
$$

Therefore the induced morphism of homology groups

$$
\left(\partial^{\prime} T+T \partial\right)_{*}: \mathrm{H}_{\bullet}(C) \longrightarrow \mathrm{H}_{\bullet}\left(C^{\prime}\right)
$$

is identically zero.

Definition 19.3.1. Let $f, g:(C, \partial) \rightarrow\left(C^{\prime}, \partial^{\prime}\right)$ be parallel chain maps. We say that $f$ and $g$ are chain-homotopic if there exists a morphism of graded groups $T: C \rightarrow C^{\prime}$ with degree 1 such that

$$
\partial^{\prime} T+T \partial=f-g .
$$

Proposition 19.3.2. Let $f, g: C \rightrightarrows C^{\prime}$ be chain-homotopic chain morphisms, then we have the equality

$$
f_{*}=g_{*}: \mathrm{H}_{\bullet}(C) \longrightarrow \mathrm{H}_{\bullet}\left(C^{\prime}\right)
$$

of the induced morphisms of homology groups.
Proof. Let $T: C \rightarrow C^{\prime}$ be a chain-homotopy between $f$ and $g$, thus $f-g=\partial^{\prime} T-T \partial$ and we have

$$
f_{*}-g_{*}=(f-g)_{*}=\left(\partial^{\prime} T-T \partial\right)_{*}=0
$$

Theorem 19.3.3. Let $f, g: X \rightrightarrows Y$ be homotopic topological morphisms. Then we have the equality

$$
f_{*}=g_{*}: \mathrm{H}_{\bullet}(X) \longrightarrow \mathrm{H}_{\bullet}(Y)
$$

of the induced morphisms of homology groups.
Proof. In view of Proposition 19.3.2 we shall merely prove that the maps $f_{\#}$ and $g_{\#}$ are chain-homotopic. Let $\eta: f \Rightarrow g$ be a homotopy and consider the inclusions $i_{0}, i_{1}: X \rightrightarrows$ $X \times I$ mapping, respectively $x \mapsto(x, 0)$ and $x \mapsto(x, 1)$.

We now discuss why it is sufficient for $i_{0 \#}$ and $i_{1 \#}$ to be chain-homotopic in order for $f_{\#}$ and $g_{\#}$ to be chain-homotopic. Suppose that $i_{0 \#}$ and $i_{1 \#}$ are indeed chainhomotopic and let $T: \operatorname{Sing}_{*} X \rightarrow \operatorname{Sing}_{*}(X \times I)$ be a degree 1 morphism of graded abelian groups such that $\partial_{X \times I} T+T \partial_{X}=i_{0 \#}-i_{1 \#}$. If we now consider the induced $\operatorname{map} \eta_{\#}: \operatorname{Sing}_{*}(X \times I) \rightarrow \operatorname{Sing}_{*} X$ we see that

$$
\begin{aligned}
\eta_{\#}\left(\partial_{X \times I} T+T \partial_{X}\right) & =\eta_{\#}\left(i_{0 \#}-i_{1 \#}\right) \\
\partial_{X}\left(\eta_{\#} T\right)+\left(\eta_{\#} T\right) \partial_{X} & =f_{\#}-g_{\#} .
\end{aligned}
$$

Thus we see that if $i_{0 \#}$ and $i_{1 \#}$ are chain-homotopic it follows that $f_{\#}$ and $g_{\#}$ are also chain-homotopic, as wanted.

We shall prove the existence of the chain-homotopy T: Sing. $X \rightarrow$ Sing. $(X \times I)$ via induction. Let $X$ denote any topological space and suppose there exists $n>0$ such that for any integer $j<n$ there exists a morphism of abelian groups $\left(T_{X}\right)_{j}: \operatorname{Sing}_{j} X \rightarrow$ $\operatorname{Sing}_{j+1}(X \times I)$ such that

$$
\begin{equation*}
\left(\partial_{X \times I}\right)_{j+1}\left(T_{X}\right)_{j}+\left(T_{X}\right)_{j-1}\left(\partial_{X}\right)_{j}=\left(i_{0 \#}\right)_{j}-\left(i_{1 \#}\right)_{j} \tag{19.1}
\end{equation*}
$$

and that for any topological morphism $h: X \rightarrow Z$ the following diagram commutes for any $j<n$ :

$$
\begin{align*}
& \operatorname{Sing}_{j} X \xrightarrow{\left(T_{X}\right)_{j}} \operatorname{Sing}_{j+1}(X \times I) \\
& \stackrel{\downarrow}{h_{\#} \downarrow} \downarrow  \tag{19.2}\\
& \operatorname{Sing}_{j} Z \xrightarrow[\left(T_{Z}\right)_{j}]{ } \operatorname{Sing}_{j+1}(\mathrm{Z} \times I)
\end{align*}
$$

We shall now define a map $\left(T_{X}\right)_{n}: \operatorname{Sing}_{n} X \rightarrow \operatorname{Sing}_{n+1}(X \times I)$. Let $\phi \in \operatorname{Sing}_{n} X$ be any singular $n$-simplex and consider the identity singular $n$-simplex id ${ }_{n} \in \operatorname{Sing}_{n} \Delta_{\text {top }}^{n}$. Notice that since $\phi_{\#}\left(\mathrm{id}_{n}\right)=\phi \mathrm{id}_{n}=\phi$, then if we define $\left(T_{\Delta_{\text {top }}^{n}}^{n}\right)_{n}: \operatorname{Sing}_{n} \Delta_{\text {top }}^{n} \rightarrow \operatorname{Sing}_{n+1}\left(\Delta_{\text {top }}^{n} \times I\right)$ following the naturality requirement presented in Eq. (19.2) then we may ask for

$$
\begin{equation*}
T_{X} \phi=T_{X} \phi_{\#} \operatorname{id}_{n}=\left(\phi \times \mathrm{id}_{I}\right)_{\#}\left(T_{\Delta_{\text {top }}^{n}} \operatorname{id}_{n}\right) \tag{19.3}
\end{equation*}
$$

Hence in order to define $\left(T_{X}\right)_{n}$ it is sufficient to define $\left(T_{\Delta_{\text {top }}^{n}}\right)_{n}$.
For the sake of convenience, we shall adopt the notation $\partial:=\partial_{\Delta_{\text {top }}^{n}}$ and $\partial^{\prime}:=\partial_{\Delta_{\text {top }}^{n} \times I}$. Let $d \in \operatorname{Sing}_{n} \Delta_{\text {top }}^{n}$ be any singular $n$-simplex and consider the chain

$$
c:=i_{0 \#} d-i_{1 \#} d-T_{\Delta_{\text {top }}^{n}} \partial_{n} d \in \operatorname{Sing}_{n}\left(\Delta_{\mathrm{top}}^{n} \times I\right)
$$

Notice that $c$ is in fact a cycle, since

$$
\begin{aligned}
\partial_{n}^{\prime} c & =\partial_{n}^{\prime} i_{0 \#} d-\partial_{n}^{\prime} i_{1 \#} d-\partial_{n}^{\prime} T_{\Delta_{\text {top }}^{n}}\left(\partial_{n} d\right) \\
& =i_{0 \#} \partial_{n} d-i_{1 \#} \partial_{n} d-\left[i_{0 \#} \partial_{n} d-i_{1 \#} \partial_{n} d-T_{\Delta_{\text {top }}^{n}} \partial_{n-1} \partial_{n} d\right] \\
& =0
\end{aligned}
$$

where we used Eq. (19.1) in order to expand the $\partial_{n}^{\prime}\left(T_{\Delta_{\text {top }}^{n}}\right)_{n-1}\left(\partial_{n} d\right)$ term. Hence we conclude that $c \in \mathrm{Z}_{n}\left(\Delta_{\text {top }}^{n} \times I\right)$. Moreover, from Theorem 19.2.3 we see that $\mathrm{H}_{n}\left(\Delta_{\text {top }}^{n} \times I\right)=$ 0 thus $c \in B_{n}\left(\Delta_{\text {top }}^{n} \times I\right)$ and there exists $b \in \operatorname{Sing}_{n+1}\left(\Delta_{\text {top }}^{n} \times I\right)$ such that $\partial_{n+1}^{\prime} b=c$. We shall define

$$
T_{\Delta_{\text {top }}^{n}} d:=b
$$

Therefore $\partial_{n+1}^{\prime} T_{\Delta_{\text {top }}^{n}} d=\partial_{n+1}^{\prime} b=c$ and hence

$$
\partial_{n+1}^{\prime} T_{\Delta_{\text {top }}^{n}} d+T_{\Delta_{\text {top }}^{n}} \partial_{n} d=i_{0 \#} d-i_{1 \#} d .
$$

This finishes the final inductive step and proves that Eq. (19.3) can be constructed:

$$
T_{X}: \operatorname{Sing}_{n} X \longrightarrow \operatorname{Sing}_{n+1}(X \times I)
$$

From the naturality expressed in Eq. (19.2) we find that for any $\phi \in \operatorname{Sing}_{n} X$ the maps $i_{0 \#}$ and $i_{1 \text { \# }}$ induce

$$
\begin{aligned}
& i_{0 \#} \phi=i_{0 \#} \phi_{\#} \mathrm{id}_{n}=\left(\phi \times \mathrm{id}_{I}\right) i_{0 \#} \mathrm{id}_{n} \\
& i_{1 \#}=i_{1 \#} \phi_{\#} \mathrm{id}_{n}=\left(\phi \times \mathrm{id}_{I}\right) i_{1 \#} \mathrm{id}_{n}
\end{aligned}
$$

Moreover, notice that

$$
\begin{aligned}
\partial_{X \times I} T_{X} \phi+T_{X} \partial_{X} \phi & =\partial_{X \times I} T_{X} \phi_{\#} \mathrm{id}_{n}+T_{X} \partial_{X} \phi_{\#} \mathrm{id}_{n} \\
& =\partial_{X \times I}\left(\phi \times \mathrm{id}_{I}\right)_{\#} T_{X} \mathrm{id}_{n}+T_{X} \phi_{\#} \partial_{\Delta_{\text {top }}^{n}} \mathrm{id}_{n} \\
& =\left(\phi \times \mathrm{id}_{I}\right)_{\#} \partial_{\Delta_{\text {top }}^{n} \times I} T_{\Delta_{\text {top }}^{n}} \mathrm{id}_{n}+\left(\phi \times \mathrm{id}_{I}\right) T_{\#} T_{\Delta_{\text {top }}^{n}} \partial_{\Delta_{\text {top }}^{n}} \mathrm{id}_{n} \\
& =\left(\phi \times \mathrm{id}_{I}\right)_{\#}\left(\partial_{\Delta_{\text {top }}^{n} \times I} T_{\Delta_{\text {top }}^{n}} \mathrm{id}_{n}+T_{\Delta_{\text {top }}^{n}} \partial_{\Delta_{\text {top }}^{n}} \mathrm{id}_{n}\right) \\
& \left.=\left(\phi \times \mathrm{id}_{I}\right) \not\right)_{\#}\left(i_{0 \#} \mathrm{id}_{n}-i_{1 \#} \mathrm{id}_{n}\right) \\
& =i_{0 \#} \phi-i_{1 \#} \phi .
\end{aligned}
$$

Moreover, naturality can be proved as follows:

$$
\begin{aligned}
T_{X} \partial_{X} \phi & =T_{X} \partial_{X} \phi_{\#} \mathrm{id}_{n} \\
& =T_{X} \phi_{\#} \partial_{\Delta_{\text {top }}^{n}} \mathrm{id}_{n} \\
& =\left(\phi \times \mathrm{id}_{I}\right) T_{\Delta_{\text {top }}^{n}} \partial_{\Delta_{\text {top }}^{n}} \mathrm{id}_{n} \\
& =\partial_{X \times I}\left(\phi \times \mathrm{id}_{I}\right)_{\#} T_{\Delta_{\text {top }}^{n}} \mathrm{id}_{n} \\
& =\partial_{X \times I} T_{X} \phi_{\#} \mathrm{id}_{n} \\
& =\partial_{X \times I} T_{X} \phi .
\end{aligned}
$$

This finishes the proof that $T_{X}$ is indeed a chain-homotopy between $i_{0 \#}$ and $i_{1 \#}$ as wanted.

### 19.4 Homotopy Invariance

Corollary 19.4.1. Let $f: X \xrightarrow{\leftrightharpoons}$ h be a homotopy equivalence, then the induced homology $\operatorname{map} f_{*}: \mathrm{H}_{\bullet}(X) \rightarrow \mathrm{H}_{\bullet}(Y)$ is an isomorphism of Z-graded abelian groups:

$$
\mathrm{H}_{\bullet}(X) \simeq \mathrm{H}_{\bullet}(Y)
$$

Proof. Let $g: Y \rightarrow X$ be a homotopy inverse of $f$. Thus $g f \sim_{h} \mathrm{id}_{X}$ and $f g \sim_{h} \mathrm{id}_{Y}$, thus via Theorem 19.3.3 we obtain $(g f)_{*}=g_{*} f_{*}=$ idsing. $^{x}$ and $(f g)_{*}=f_{*} g_{*}=\operatorname{id}_{\text {Sing. }} \gamma$. This shows that $f_{*}$ and $g_{*}$ are inverses of each other and thus $f_{*}$ is an isomorphism.

Corollary 19.4.2. Let $A$ be a retract of $X$ and $t: A \hookrightarrow X$ be the canonical inclusion, then $\iota_{*}: \mathrm{H}_{\bullet}(A) \rightarrow \mathrm{H}_{\bullet}(X)$ is a split monomorphism onto a direct summand. If $A$ is a deformation retract of $X$, then $\iota_{*}$ is an isomorphism

Proof. Since $\iota$ admits a left inverse $r: X \rightarrow A$ then $r_{*} \iota_{*}=\mathrm{id}_{\mathrm{H}_{\bullet}(A)}$ shows that $\iota_{*}$ is a split monomorphism. We now show that $\iota_{*}$ maps $\mathrm{H}_{\bullet}(A)$ onto a direct summand of $\mathrm{H}_{\bullet}(X)$. Let $G_{1}:=\operatorname{im} \iota_{*}$ and $G_{2}:=\operatorname{ker} r_{*}$ be two subgroups of $H_{\bullet}(X)$. Let $\alpha \in H_{\bullet}(X)$ be any element and notice that we can write $\alpha=\iota_{*} r_{*} \alpha+\left(\alpha-\iota_{*} r_{*} \alpha\right)$, where $\iota_{*} r_{*} \alpha \in \operatorname{im} \iota_{*}$ and since $r_{*}$ is a left inverse of $\iota_{*}$ then $\alpha-\iota_{*} r_{*} \alpha \in \operatorname{ker} r_{*}$. Therefore we have $\mathrm{H}_{\bullet}(X)=G_{1}+G_{2}$, furthermore, if $\beta \in G_{1} \cap G_{2}$ we have that there exists $\gamma \in \mathrm{H}_{\bullet}(A)$ such that $\iota_{*} \gamma=\beta$ and $r_{*} \beta=0$, therefore

$$
\gamma=r_{*} i_{*} \gamma=r_{*} \beta=0
$$

hence $\beta=0$. This shows that $G_{1} \cap G_{2}=0$, thus $G_{1}+G_{2}=G_{1} \oplus G_{2}$ and $H_{\bullet}(X)=$ $G_{1} \oplus G_{2}$.

### 19.5 Long Exact Sequence Theorem

Theorem 19.5.1. Given a short exact sequence of chain complexes with degree zero chain maps:

$$
0 \longrightarrow A \ggg 0
$$

the following long sequence of homology groups is exact:

$$
\cdots \xrightarrow{i_{*}} \mathrm{H}_{p}(B) \xrightarrow{j_{*}} \mathrm{H}_{p}(C) \xrightarrow{\delta} \mathrm{H}_{p-1}(A) \xrightarrow{i_{*}} \cdots
$$

Proof. We shall define $\delta: \mathrm{H}_{\bullet}(C) \rightarrow \mathrm{H}_{\bullet}(A)$ with degree -1 . Let $c \in \mathrm{Z}_{p}(C)$ and since $j_{p}$ is surjective, there exists $b \in B_{p}$ such that $j_{p} b=c$. Furthermore,

$$
j_{p-1} \partial_{p}^{B} b=\partial_{p}^{C} j_{p} b=\partial_{p}^{C} c=0
$$

from the fact that $j$ is a chain map-therefore $\partial_{p}^{B} b \in \operatorname{ker} j_{p-1}=\operatorname{im} i_{p-1}$. Since $\partial_{p}^{B} b \in$ $\operatorname{im} i_{p-1}$ then there exists a unique $a \in A_{p-1}$ such that $i_{p-1} a=\partial_{p}^{B} b$-the uniqueness of $a$ comes from the fact that $i$ is an injective map. Define the map $\delta_{p}: \mathrm{H}_{p}(C) \rightarrow \mathrm{H}_{p-1}(A)$ to take $\delta_{p}[c]:=[a]$. We need to show that $\delta_{p}$ is well defined, which requires us to prove that the mapping $[c] \mapsto[a]$ is independent of the choice of $b$ and $c$ :

- Let $b^{\prime} \in B_{p}$ be such that $j_{p} b^{\prime}=c$, thus $b^{\prime}-b \in \operatorname{ker} j_{p}$ and since ker $j_{p}=\operatorname{im} i_{p}$ it follows that there exists $a^{\prime} \in A_{p}$ for which $i_{p} a^{\prime}=b^{\prime}-b$. Therefore

$$
\partial_{p}^{B} b^{\prime}=\partial_{p}^{B} b+\partial_{p}^{B} i_{p} a^{\prime}=i_{p} a+i_{p} \partial_{p}^{A} a^{\prime}=i_{p}\left(a+\partial_{p}^{A} a^{\prime}\right)
$$

which shows that $\left[a+\partial_{p}^{A} a^{\prime}\right]=[a]$, therefore the construction is independent of the choice of $b$.

- Let $\left[c^{\prime}\right]=[c]$ so that $c^{\prime}-c \in B_{p}(C)$, and choose $\alpha \in C_{p+1}$ such that $c^{\prime}-c=\partial_{p+1}^{C} \alpha$. By the surjectiveness of $j$, choose $b^{\prime} \in B_{p+1}$ for which $\alpha=j_{p+1} b^{\prime}$. Therefore

$$
c^{\prime}=c+\partial_{p+1}^{C} \alpha=c+\partial_{p+1}^{C} j_{p+1} b^{\prime}=j_{p}(b)+j_{p}\left(\partial_{p+1}^{B} b^{\prime}\right)=j_{p}\left(b+\partial_{p+1}^{B} b^{\prime}\right) .
$$

Moreover, one also has

$$
\partial_{p}^{B}\left(b+\partial_{p+1}^{B} b^{\prime}\right)=\partial_{p}^{B} b+\partial_{p}^{B} \partial_{p+1}^{B} b^{\prime}=\partial_{p}^{B} b,
$$

therefore $i_{p-1} a=\partial_{p}^{B} b=\partial_{p}^{B}\left(b+\partial_{p+1}^{B} b^{\prime}\right)$. This shows us that the construction is agnostic of the choice of representative of the class [ $c$ ].

It remains for us to prove that the resulting sequence of homology groups is exact:

- $\left(\operatorname{im} i_{*}=\operatorname{ker} j_{*}\right)$ Notice that by functoriality we have $j_{*} i_{*}=(j i)_{*}=0$ since im $i \subseteq$ $\operatorname{ker} j$-this shows that $\operatorname{im} i_{*} \subseteq \operatorname{ker} j_{*}$. On the other hand, if $[b] \in \operatorname{ker} j_{* p}$ is any element, then $j_{p} b \in B_{p}(C)$ and there exists $c \in C_{p+1}$ such that $j_{p} b=\partial_{p+1}^{C} c$. Since $j$ is surjective, there also exists $b^{\prime} \in B_{p+1}$ such that $j_{p+1} b^{\prime}=c$. Notice that

$$
j_{p} \partial_{p+1}^{B} b^{\prime}=\partial_{p}^{B} j_{p+1} b^{\prime}=\partial_{p}^{B} c=j_{p} b,
$$

therefore $b-\partial_{p+1}^{B} b^{\prime} \in \operatorname{ker} j_{p}=\operatorname{im} i_{p}$. Since $i$ is injective it follows that there exists a unique $a \in A_{p}$ such that $i a=b-\partial_{p+1}^{B} b^{\prime}$. Notice that since $[b] \in \mathrm{H}_{p}(B)$ then in particular $b \in \mathrm{Z}_{p}(B)$, therefore

$$
i_{p-1} \partial_{p}^{A} a=\partial_{p}^{B} i_{p} a=\partial_{p}^{B}\left(b-\partial_{p+1}^{B} b^{\prime}\right)=\partial_{p}^{B} b-\partial_{p}^{B} \partial_{p+1}^{B} b^{\prime}=0 .
$$

From the injectivity of $i$ we conclude that $\partial_{p}^{A} a=0$, therefore $a \in Z_{p}(A)$ and hence

$$
[b]=\left[b-\partial_{p+1}^{B} b^{\prime}\right]=\left[i_{p} a\right]=i_{* p}[a]
$$

which proves that $\operatorname{ker} j_{* p} \subseteq \operatorname{im} i_{* p}$ and in general ker $j_{*} \subseteq \operatorname{im} i_{*}$.

- (im $\left.j_{*}=\operatorname{ker} \delta\right)$ By the definition of $\delta$ we have $\delta j_{*}=0$ thus $\operatorname{im} j_{*} \subseteq \operatorname{ker} \delta$. Let $[c] \in \operatorname{ker} \delta_{p}$ then $\delta_{p}[c]=[a]=0$, which means that $a \in \mathrm{~B}_{p-1}(A)$ and therefore there exists $a^{\prime} \in A_{p}$ such that $a=\partial_{p}^{A} a^{\prime}$. Since $j$ is surjective, let $b \in B_{p}$ be such that $j_{p} b=c$, then

$$
j_{p}\left(b-i_{p} a^{\prime}\right)=j_{p} b-j_{p} i_{p} a^{\prime}=j_{p} b=c
$$

moreover we also have

$$
\partial_{p}^{B}\left(b-i_{p} a^{\prime}\right)=\partial_{p}^{B} b-\partial_{p}^{B} i_{p} a^{\prime}=\partial_{p}^{B} b-i_{p-1} \partial_{p}^{A} a^{\prime}=\partial_{p}^{B} b-i_{p-1} a=0
$$

therefore $b-i_{p} a^{\prime} \in Z_{p}(B)$ and hence

$$
j_{* p}\left[b-i_{p} a^{\prime}\right]=\left[j_{p}\left(b-i_{p} a^{\prime}\right)\right]=[c]
$$

showing that ker $\delta \subseteq \operatorname{im} j_{*}$.

- $\left(\operatorname{im} \delta=\operatorname{ker} i_{*}\right)$ By the definition of $\delta$, if $a \in \operatorname{im} \delta_{p}$ with $\delta[c]=[a]$, and both $j_{p} b=c$ and $i_{p-1} a=\partial_{p}^{B} b$ then

$$
i_{* p-1}[a]=\left[i_{p-1} a\right]=\left[\partial_{p}^{B} b\right]=0
$$

thus $\operatorname{im} \delta \subseteq \operatorname{ker} i_{*}$. Moreover, if $\left[a^{\prime}\right] \in \operatorname{ker} i_{* p-1}$, let $i_{p-1} a=\partial_{p}^{B} b$ for some $b \in B_{p}$. Since

$$
\partial_{p}^{C} j_{p} b=j_{p-1} \partial_{p}^{B} b=j_{p-1} i_{p-1} a=0
$$

then $j_{p} b \in Z_{p}(C)$ and hence $\delta[j b]=[a]$. Therefore ker $i_{*} \subseteq \operatorname{im} \delta$.

Proposition 19.5.2. The long exact sequence induced by short exact sequences of chain complexes is natural. In other terms, given a commutative diagram with exact rows:

then the following diagram with long exact rows commutes:


### 19.6 Reduced Homology

Definition 19.6.1 (Reduced homology). Let $X$ be a non-empty topological space, and define a surjective morphism $\varepsilon: \operatorname{Sing}_{0} X \rightarrow \mathbf{Z}$ given by $\sum_{j} n_{j} \phi_{j} \mapsto \sum_{j} n_{j}$. Notice that $\operatorname{im} \partial_{1} \subseteq \operatorname{ker} \varepsilon$ since given any $\tau \in \operatorname{Sing}_{1} X$ we have

$$
\varepsilon \partial_{1} \tau=\varepsilon(\tau(1)-\tau(0))=1-1=0
$$

Therefore the sequence

$$
\cdots \longrightarrow \operatorname{Sing}_{2} X \xrightarrow{\partial_{2}} \operatorname{Sing}_{1} X \xrightarrow{\partial_{1}} \operatorname{Sing}_{0} X \xrightarrow{\varepsilon} \mathbf{Z} \longrightarrow 0
$$

is a chain complex-in fact, we call such complex the augmented chain complex of $X$. It is to be noticed that from the isomorphism theorem for groups there exists a unique morphism $\varepsilon_{*}: \mathrm{H}_{0}(X) \rightarrow \mathbf{Z}$ such that the following diagram commutes


If we now define a Z-graded group $\widetilde{H}_{\bullet}(X)$ as follows: for each $p>0$ let $\widetilde{\mathrm{H}}_{p}(X):=\mathrm{H}_{p}(X)$, while $\widetilde{\mathrm{H}}_{0}(X):=\operatorname{ker} \varepsilon / \operatorname{im} \partial_{1}$. The group $\widetilde{\mathrm{H}}_{p}(X)$ is called the $p$-th reduced homology group of $X$.

In particular we know that $\operatorname{ker} \varepsilon_{*}=\operatorname{ker} \varepsilon / \operatorname{im} \partial_{1}$, therefore we may define a choicedependent isomorphism

$$
\widetilde{\mathrm{H}}_{0}(X) \oplus \mathbf{Z} \simeq \mathrm{H}_{0}(X)
$$

as follows: given $\psi \in \operatorname{Sing}_{0} X$, define the mappings $\left(\left[\sum_{j} n_{j} \phi_{j}\right], m\right) \mapsto\left[\sum_{j} n_{j} \phi_{j}+m \psi\right]$ together with an inverse $\left[\sum_{j} n_{j} \phi_{j}\right] \mapsto\left(\left[\sum_{j} n_{j} \phi_{j}-\sum_{j} n_{j} \psi\right], \sum_{j} n_{j}\right)$. Another way to see the existence of this isomorphism is to notice that since $\operatorname{im} \varepsilon_{*}=\mathbf{Z}$ and ker $\varepsilon_{*}=\widetilde{\mathrm{H}}_{0}(X)$ then there exists a short exact sequence

$$
0 \longrightarrow \tilde{\mathrm{H}}_{0}(X) \longleftrightarrow \mathrm{H}_{0}(X) \xrightarrow{\varepsilon} \mathbf{Z} \longrightarrow 0
$$

and since $\mathbf{Z}$ is free then the sequence splits and there exists an isomorphism $H_{0}(X) \simeq$ $\widetilde{\mathrm{H}}_{0}(X) \oplus \mathbf{Z}$ as expected.

### 19.7 Mayer-Vietoris Sequence

## Barycentric Subdivision

## To be added

## Mayer-Vietoris Theorem

Definition 19.7.1. Let $X$ be a topological space and $\mathcal{U}$ be an open cover of $X$. Denote by Sing ${ }_{\bullet}^{\mathcal{U}} X$ the $\mathbf{Z}$-graded abelian group given by groups $\operatorname{Sing}_{n}^{\mathcal{U}} X$ composed of singular $n$-simplices ${ }^{1} \phi: \Delta_{\text {top }}^{n} \rightarrow X$ such that $\operatorname{im} \phi \subseteq U$ for some $U \in \mathcal{U}$-for each $n \in \mathbf{Z}$. Since $\operatorname{im} \partial^{j} \phi \subseteq \operatorname{im} \phi$, the boundary map

$$
\partial: \operatorname{Sing}_{\bullet}^{\mathcal{U}} X \rightarrow \operatorname{Sing}_{\bullet}^{\mathcal{U}} X
$$

is well defined and has degree -1 .
Given another topological space $Y$ together with an open covering $\mathcal{V}$ and a continuous map $f: X \rightarrow Y$ for which each $U \in \mathcal{U}$ has image $f U \subseteq V$ for some $V \in \mathcal{V}$, we have an induced chain map $f_{\#}:$ Sing $\mathcal{U}_{\bullet} X \rightarrow$ Sing. $Y$.

Theorem 19.7.2. Let $\mathcal{U}$ be a collection of subsets of $X$ such that the interior of $\mathcal{U}$ is an open cover for $X$, then the canonical inclusion map $t: \operatorname{Sing}_{\bullet}^{\mathcal{U}} X \rightarrow \operatorname{Sing}_{\bullet} X$ induces an isomorphism

$$
\iota_{*}: \mathrm{H}_{\bullet}\left(\operatorname{Sing} \mathcal{U}_{\bullet} X\right) \stackrel{\simeq}{\leftrightarrows} \mathrm{H}_{\bullet}(X) .
$$

Theorem 19.7.3. Let $X$ be a space and $U, V \subseteq X$ be subsets such that $\operatorname{Int} U \cup \operatorname{Int} V=X$ with inclusion maps:


Then there exists a long exact sequence of homology groups, called the Mayer-Vietoris sequence, of the form

$$
\cdots \longrightarrow \mathrm{H}_{n}(U \cap V) \xrightarrow{i_{*} \oplus j_{*}} \mathrm{H}_{n}(U) \oplus \mathrm{H}_{n}(V) \xrightarrow{k_{*}-\ell_{*}} \mathrm{H}_{n}(X) \xrightarrow{\delta} \mathrm{H}_{n-1}(U \cap V) \longrightarrow \cdots
$$

Moreover, this sequence is natural-that is, given subsets $U^{\prime}, V^{\prime} \subseteq X^{\prime}$ with $\operatorname{Int} U^{\prime} \cup$ Int $V^{\prime}=X$ and a continuous map $f: X \rightarrow X^{\prime}$ for which $f U \subseteq U^{\prime}$ and $f V \subseteq V^{\prime}$ then the following diagram commutes:


Proof. For convenience, let $\mathcal{U}:=\{U, V\}$. Let $S_{U}$ and $S_{V}$ be the sets consisting of all singular $n$-simplices in $U$ and $V$, respectively. If F : Set $\rightarrow \mathrm{Ab}$ denotes the free abelian group functor, then it follows that

$$
\operatorname{Sing}_{n}(U \cap V)=\mathrm{F}\left(S_{U} \cap S_{V}\right) \quad \text { and } \quad \operatorname{Sing}_{n}^{\mathcal{U}} X=\mathrm{F}\left(S_{U} \cup S_{V}\right)
$$

[^28]Define a morphism of groups

$$
h: \mathrm{F}\left(S_{U}\right) \oplus \mathrm{F}\left(S_{V}\right) \rightarrow \mathrm{F}\left(S_{U} \cup S_{V}\right)
$$

to be given by $(\phi, \psi) \mapsto \phi-\psi$, which gives an epimorphism. Further, define a monomorphism of groups

$$
g: \mathrm{F}\left(S_{U} \cap S_{V}\right) \mapsto \mathrm{F}\left(S_{U}\right) \oplus \mathrm{F}\left(S_{V}\right)
$$

mapping $\tau \mapsto(\tau, \tau)$. Notice that from construction we have $h g=0$, hence $\operatorname{im} g \subseteq \operatorname{ker} h$.
On the other hand, let $(\Phi, \Psi):=\left(\sum_{\phi \in S_{U}} n_{\phi} \phi, \sum_{\psi \in S_{V}} m_{\psi} \psi\right) \in \operatorname{ker} h$, which is the case if and only if for each $\phi \in S_{U}$ associated with a non-zero coefficient $n_{\phi}$ there exists $\psi \in S_{V}$ with $m_{\psi}=n_{\phi}$ and $\psi=\phi$-this construction should be bijective in the singular simplices that have non-zero coefficient in order to obtain $\Phi+\Psi=0$. From this consideration, we conclude that for $(\Phi, \Psi)$ to be an element of the kernel of $h$ it must be the case that each non-zero simplex lies in the intersection $S_{U} \cap S_{V}$, therefore it is clear that $\operatorname{ker} h \subseteq \operatorname{im} g$. This shows us that

$$
\begin{equation*}
0 \longrightarrow \operatorname{Sing}_{n}(U \cap V) \stackrel{g}{\longrightarrow} \operatorname{Sing}_{n} U \oplus \operatorname{Sing}_{n} V \xrightarrow{h} \operatorname{Sing}_{n}(U \cup V) \longrightarrow 0 \tag{19.4}
\end{equation*}
$$

is a short exact sequence, which can be extended as a short exact sequence of chain complexes and degree zero chain maps $g$ and $h$.

Using the long exact sequence theorem on Eq. (19.4) we obtain an exact sequence of homology groups

$$
\cdots \longrightarrow \mathrm{H}_{n}(U \cap V) \xrightarrow{g_{*}} \mathrm{H}_{n}(U) \oplus \mathrm{H}_{n}(V) \xrightarrow{h_{*}} \mathrm{H}_{n}^{\mathcal{U}}(X) \xrightarrow{\delta} \mathrm{H}_{n-1}(U \cap V) \longrightarrow \cdots
$$

Now using the isomorphism from Theorem 19.7.2 we obtain the desired long exact sequence.

Corollary 19.7.4. Let $X$ be a non-empty topological space, and $U, V \subseteq X$ be intersecting subsets such that $\operatorname{Int} U \cup \operatorname{Int} V=X$. There exists an exact sequence

$$
\cdots \rightarrow \widetilde{\mathrm{H}}_{p}(U \cap V) \rightarrow \widetilde{\mathrm{H}}_{p}(U) \oplus \widetilde{\mathrm{H}}_{p}(V) \rightarrow \widetilde{\mathrm{H}}_{p}(X) \rightarrow \widetilde{\mathrm{H}}_{p-1}(U \cap V) \rightarrow \cdots
$$

and the reduced homology makes this sequence end:

$$
\cdots \longrightarrow \widetilde{\mathrm{H}}_{0}(U) \oplus \tilde{\mathrm{H}}_{0}(V) \longrightarrow \widetilde{\mathrm{H}}_{0}(X) \longrightarrow 0
$$

## Applications of The Mayer-Vietoris Sequence

Example 19.7.5 (The circle). Consider the circle $S^{1}$ and denote by $s$ and $n$ its south and north poles, respectively. Define subsets $U:=S^{1} \backslash s$ and $V:=S^{1} \backslash n$. First let us calculate the first homology group of $S^{1}$ : to that end, we can use the Mayer-Vietoris sequence for the triple $\left(S^{1}, U, V\right)$ to obtain an exact sequence

$$
\mathrm{H}_{1}(U) \oplus \mathrm{H}_{1}(V) \xrightarrow{k_{*}-\ell_{*}} \mathrm{H}_{1}\left(S^{1}\right) \xrightarrow{\delta} \mathrm{H}_{0}(U \cap V) \xrightarrow{i_{*} \oplus j_{*}} \mathrm{H}_{0}(U) \oplus \mathrm{H}_{0}(V)
$$

Notice that both $U$ and $V$ are contractible spaces, therefore they have the same homotopy as the single point space, that is

$$
\mathrm{H}_{p}(U)=\mathrm{H}_{p}(V)= \begin{cases}\mathbf{Z}, & \text { if } p=0 \\ 0, & \text { otherwise }\end{cases}
$$

Thus $\mathrm{H}_{1}(U) \oplus \mathrm{H}_{1}(V)=0$-implying that $\delta$ is a monomorphism—and we also have $\mathrm{H}_{0}(U) \oplus \mathrm{H}_{0}(V) \simeq \mathbf{Z} \oplus \mathbf{Z}$. On the other hand, the set $U \cap V$ is homotopic to a space consisting of two points-that is, $U \cap V \simeq_{h}\{*\} \amalg\{*\}$-thus in particular we have an isomorphism $\mathrm{H}_{0}(U \cap V) \simeq \mathbf{Z} \oplus \mathbf{Z}$ by mapping $a \phi+b \psi \mapsto(a, b)$. In summary, we obtain the following exact sequence

$$
0 \longrightarrow \mathrm{H}_{1}\left(S^{1}\right) \stackrel{\delta}{\longrightarrow} \mathbf{Z} \oplus \mathbf{Z} \xrightarrow{i_{*}^{\prime} \oplus j_{*}^{\prime}} \mathbf{Z} \oplus \mathbf{Z}
$$

where both $i_{*}^{\prime}, j_{*}^{\prime}: \mathbf{Z} \oplus \mathbf{Z} \rightrightarrows \mathbf{Z}$ map $(a, b) \mapsto a+b$, since $i_{*}$ and $j_{*}$ are inclusions. Given an element $a \phi+b \psi \in \mathrm{H}_{0}(U \cap V)$, we find that $a \phi+b \psi$ corresponds to an element of the kernel of $i_{*}^{\prime} \oplus j_{*}^{\prime}$ if and only if $a=-b$. Hence $\operatorname{ker}\left(i_{*}^{\prime} \oplus j_{*}^{\prime}\right)$ is a subgroup of $\mathbf{Z} \oplus \mathbf{Z}$ generated by elements of the form $(a,-a)$-that is,

$$
\operatorname{ker}\left(i_{*}^{\prime} \oplus j_{*}^{\prime}\right)=\operatorname{im} \delta \simeq \mathbf{Z}
$$

Since $\delta$ is injective, then $\mathrm{H}_{1}\left(S^{1}\right) \simeq \operatorname{im} \delta$, therefore $\mathrm{H}_{1}\left(S^{1}\right) \simeq \mathbf{Z}$.
Moreover, for the case of $p>1$ we have an exact sequence

$$
\mathrm{H}_{p}(U) \oplus \mathrm{H}_{p}(V) \xrightarrow{k_{*}-\ell_{*}} \mathrm{H}_{p}\left(S^{1}\right) \xrightarrow{\delta} \mathrm{H}_{p-1}(U \cap V)
$$

however from previous considerations we know that for $p>1$ we have both $\mathrm{H}_{p}(U)=$ $\mathrm{H}_{p}(V)=0$ and $\mathrm{H}_{p-1}(U \cap V)=0$. This shows that ker $\delta=\mathrm{H}_{p}\left(S^{1}\right)$ and $\operatorname{since} \operatorname{im}\left(k_{*}-\ell_{*}\right)=0$ it follows from exactness that $\mathrm{H}_{p}\left(S^{1}\right)=0$. This results shows us that:

$$
\mathrm{H}_{p}\left(S^{1}\right)= \begin{cases}\mathbf{Z}, & \text { if } p=1 \\ 0, & \text { otherwise }\end{cases}
$$

Example 19.7.6 ( $n$-spheres). Let's consider the general case of an $n$-sphere $S^{n} \subseteq \mathbf{R}^{n+1}$, where $n>0$. Let $n, s \in S^{n}$ be the north and south poles, respectively, and define subsets $U:=S^{n} \backslash n$ and $V:=S^{n} \backslash s$. Via the stereographic projection we know that there are topological isomorphisms

$$
U \simeq \mathbf{R}^{n} \simeq V \quad \text { and } \quad U \cap V \simeq \mathbf{R}^{n} \backslash 0
$$

furthermore, $S^{n-1}$ is a deformation retract of $\mathbf{R}^{n} \backslash 0$, hence $U \cap V \simeq_{h} S^{n-1}$. Therefore we have, for any $p$ :


Since $\mathbf{R}^{n}$ is contractible then for $p>0$ we have $\mathrm{H}_{p}\left(\mathbf{R}^{n}\right)=0$, therefore both end terms of the above sequence are zero. By exactness of the sequence, this implies that for $p>1$ the map $\delta$ is an isomorphism $\mathrm{H}_{p}\left(S^{n}\right) \simeq \mathrm{H}_{p-1}\left(S^{n-1}\right)$. Therefore, for $p=n$ we find recursively that $\mathrm{H}_{n}\left(S^{n}\right) \simeq \mathrm{H}_{1}\left(S^{1}\right) \simeq \mathbf{Z}$. In the case of $p=1$ and $n>1$ we have that both $d_{*}$ and $g_{*}$ are monomorphisms, thus $\mathrm{H}_{1}\left(S^{n}\right) \simeq 0$-this also follows from Hurewicz's theorem. Otherwise, if $p>1$ and $p \neq n$ then $\mathrm{H}_{p}\left(S^{n}\right)$. This settles the following: for $n>0$

$$
\mathrm{H}_{p}\left(S^{n}\right)= \begin{cases}\mathbf{Z}, & \text { if } p \in\{0, n\} \\ 0, & \text { otherwise }\end{cases}
$$

for the case $n=0, S^{0}$ consists of two points-therefore has two path connected components, hence $\mathrm{H}_{0}\left(S^{0}\right) \simeq \mathbf{Z} \oplus \mathbf{Z}$, thus

$$
\mathrm{H}_{p}\left(S^{0}\right)= \begin{cases}\mathbf{Z} \oplus \mathbf{Z}, & \text { if } p=0 \\ 0, & \text { otherwise }\end{cases}
$$

Example 19.7.7 (Torus). Consider the torus $T$. We shall calculate the homology of $T$ using two different coverings $U$ and $V$ :
(a) Let $U$ and $V$ be intersecting subsets of $T$, where $U$ is one half of the torus and $V$ is the other half-therefore:

$$
U \simeq V \simeq_{h} S^{1} \quad \text { and } \quad U \cap V=S^{1} \amalg S^{1} .
$$

Since $T$ is path-connected, we know that $\mathrm{H}_{0}(T)=\mathbf{Z}$.

- For the case where $p>2$ we have

hence from exactness we obtain $\mathrm{H}_{p}(T)=0$.
- For the case $p=2$ we have


Since the sequence is exact, then $\mathrm{H}_{2}(T) \mapsto \mathbf{Z} \oplus \mathbf{Z}$ is a monomorphism. Let's analyse the map $i_{*} \oplus j_{*}$ : let $u, v \in Z_{1}(U \cap V)$ be equators, where $u$ sits in one of the cylinders while $v$ in the other. Notice however that

$$
\left(i_{*} \oplus j_{*}\right)[\alpha]_{U \cap V}=(\alpha, \beta)=\left(i_{*} \oplus j_{*}\right)[\beta]_{U \cap V}
$$

since the classes of $\alpha$ and $\beta$ are equal in both $\mathrm{H}_{1}(U)$ and $\mathrm{H}_{1}(V)$. From injectivity of $\delta$ we know that $\mathrm{H}_{2}(T) \simeq \operatorname{im} \delta$, moreover exactness implies in $\operatorname{im} \delta=\operatorname{ker}\left(i_{*} \oplus\right.$ $j_{*}$ ). Notice that

$$
\operatorname{ker}\left(i_{*} \oplus j_{*}\right)=\mathbf{Z}[\alpha-\beta]_{U \cap V} \simeq \mathbf{Z},
$$

therefore $\mathrm{H}_{2}(T) \simeq \mathbf{Z}$.

- For the case where $p=1$ we have


In order to analyse the map $\left(i_{*} \oplus j_{*}\right)_{0}$ we can take distinct points $u, v \in U \cap V-$ each contained in one of the two connected components of $U \cap V$-so that [ $u]_{U}$ generates $\mathrm{H}_{0}(U)$ while $[v]_{V}$ generates $\mathrm{H}_{0}(V)$. It follows that

$$
\mathrm{H}_{0}(U \cap V) \simeq \mathbf{Z}[u]_{U \cap V} \oplus \mathbf{Z}[v]_{U \cap V} .
$$

Notice that since $[u]_{U}=[v]_{U}$ and $[u]_{V}=[v]_{V}$ then

$$
\left(i_{*} \oplus j_{*}\right)_{0}[u]_{U \cap V}=\left(i_{*} \oplus j_{*}\right)_{0}[v]_{U \cap V},
$$

therefore, $\operatorname{ker}\left(i_{*} \oplus j_{*}\right)_{0}=\mathbf{Z}[u-v]_{U \cap V} \simeq \mathbf{Z}$ and from exactness we have

$$
\operatorname{im} \delta=\operatorname{ker}\left(i_{*} \oplus j_{*}\right)_{0} \simeq \mathbf{Z} .
$$

It should be noted that the long sequence can be split into short exact sequences, for instance, we are interested in the following induced short sequence:


Moreover, we have

$$
\operatorname{im}\left(k_{*}-\ell_{*}\right) \simeq \frac{\mathbf{Z} \oplus \mathbf{Z}}{\operatorname{ker}\left(k_{*}-\ell_{*}\right)} \simeq \frac{\mathbf{Z} \oplus \mathbf{Z}}{\operatorname{im}\left(i_{*} \oplus j_{*}\right)_{1}} \simeq \frac{\mathbf{Z} \oplus \mathbf{Z}}{\mathbf{Z}} \simeq \mathbf{Z},
$$

therefore the short exact sequence above becomes the split sequence

$$
0 \longrightarrow \mathbf{Z} \longrightarrow \mathrm{H}_{1}(T) \longrightarrow \mathbf{Z} \longrightarrow 0
$$

which implies in $\mathrm{H}_{1}(T) \simeq \mathbf{Z} \oplus \mathbf{Z}$.

This can be summarised as

$$
\mathrm{H}_{p}(T)= \begin{cases}\mathbf{Z}, & \text { if } p=0 \\ \mathbf{Z} \oplus \mathbf{Z}, & \text { if } p=1 \\ \mathbf{Z}, & \text { if } p=2 \\ 0, & \text { otherwise }\end{cases}
$$

(b) Let $p \in T$ be any point. Consider the cover composed of $U:=T \backslash p$ and $V$ being a disk around $p$. Notice that $V$ is topologically isomorphic to a disk in $\mathbf{R}^{2}$, which is a convex euclidean set, thus

$$
\mathrm{H}_{p}(V)= \begin{cases}\mathbf{Z}, & \text { if } p=0 \\ 0, & \text { otherwise }\end{cases}
$$

Also notice that $U \cap V=V \backslash p$, which is isomorphic to $S^{1}$, hence $\mathrm{H}_{p}(U \cap V) \simeq$ $\mathrm{H}_{p}\left(S^{1}\right)$. The reader should now draw for himself the gluing diagram of $T$ without a point, noticing that we can homotopically remove the inside face of the diagram, remaining only the gluing boundaries, which after being glued we see that $T \backslash p \simeq_{h}$ $S^{1} \vee S^{1}$, hence $\mathrm{H}_{p}(U) \simeq \mathrm{H}_{p}\left(S^{1} \vee S^{1}\right)$. Using the the Hurewicz's theorem we obtain:

$$
\mathrm{H}_{1}(U) \simeq \mathrm{H}_{1}\left(S^{1} \vee S^{1}\right) \simeq \pi_{1}^{\mathrm{Ab}}\left(S^{1} \vee S^{1}\right)=(\mathbf{Z} * \mathbf{Z})^{\mathrm{Ab}} \simeq \mathbf{Z} \oplus \mathbf{Z}
$$

Notice that for $p \geqslant 3$ we have

$$
0=\mathrm{H}_{p}(U) \oplus \mathrm{H}_{p}(V) \longrightarrow \mathrm{H}_{p}(T) \longrightarrow \mathrm{H}_{p-1}(U \cap V)=0
$$

therefore from exactness we get $\mathrm{H}_{p}(T)=0$. For the case where $p=2$ we have:


Since $H_{1}(V)$ is zero, then in particular the map $\mathrm{H}_{1}(U \cap V) \rightarrow \mathrm{H}_{1}(U) \oplus \mathrm{H}_{1}(V)$ sends $[\phi] \mapsto(0,0)$, therefore the induced map $\mathbf{Z} \rightarrow \mathbf{Z} \oplus \mathbf{Z}$ is zero, that is, has kernel $\mathbf{Z}$. Since the sequence is exact, the image of $\mathrm{H}_{2}(T) \rightarrow \mathbf{Z}$ equals $\mathbf{Z}$, therefore being both an injection and surjection, proving that $\mathrm{H}_{2}(T) \simeq \mathbf{Z}$.
Now for $p=1$ :


Furthermore, from exactness we know that $\mathbf{Z} \oplus \mathbf{Z} \rightarrow \mathrm{H}_{1}(T)$ has null kernel, hence a monomorphism, since the image of $\mathbf{Z} \rightarrow \mathbf{Z} \oplus \mathbf{Z}$ is zero. On the other hand, since $\mathrm{H}_{0}(U \cap V) \rightarrow \mathrm{H}_{0}(U) \oplus \mathrm{H}_{0}(V)$ induces an injective mapping $1 \mapsto(1,1)$ of the form $\mathbf{Z} \rightarrow \mathbf{Z} \oplus \mathbf{Z}$, then by exactness it follows that the image of $\mathrm{H}_{1}(T) \rightarrow \mathrm{H}_{0}(U \cap V)$ is identically zero. This proves that $\mathbf{Z} \oplus \mathbf{Z} \rightarrow \mathrm{H}_{1}(T)$ is also an epimorphism, which shows the existence of an isomorphism $\mathrm{H}_{1}(T) \simeq \mathbf{Z} \oplus \mathbf{Z}$. In summary we have:

$$
\mathrm{H}_{p}(T)= \begin{cases}\mathbf{Z}, & \text { if } p=0 \\ \mathbf{Z} \oplus \mathbf{Z}, & \text { if } p=1 \\ \mathbf{Z}, & \text { if } p=2 \\ 0, & \text { otherwise }\end{cases}
$$

where $\mathrm{H}_{0}(T)=\mathbf{Z}$ comes from the fact that $T$ is path-connected.
Example 19.7.8 (Klein bottle). Let $K$ denote the Klein bottle and let both $M$ and $M^{\prime}$ denote copies of a Möbius band. If the relation $\sim$ denoted the gluing of $M$ and $M^{\prime}$ along their boundary, then $K \simeq\left(M \amalg M^{\prime}\right) / \sim$. Moreover, we know that $M=M^{\prime} \simeq_{h} S^{1}$ and $M \cap M^{\prime} \simeq_{h} S^{1}$. From this we know that for each $p>2$ we have

therefore $\mathrm{H}_{p}(K)=0$ from exactness. Let's consider the case where $p=2$ :


Consider the circle $\sigma \in Z_{1}\left(M \cap M^{\prime}\right)$, which generates all 1-cycles of the intersection, then when mapped back to either $M$ or $M^{\prime}$ we obtain a cycle looping two times around the boundary of the strips, therefore the induced mapping $\mathbf{Z} \rightarrow \mathbf{Z} \oplus \mathbf{Z}$ takes $1 \mapsto(2,2)$, showing that the map is injective-hence has a null kernel. Since $\mathrm{H}_{2}(\mathrm{~K}) \mapsto \mathbf{Z}$ is also injective, then $\mathrm{H}_{2}(\mathrm{~K})$ is isomorphic to its image in $\mathbf{Z}$-which on the other hand is equal to the kernel of $\mathbf{Z} \rightarrow \mathbf{Z} \oplus \mathbf{Z}$ by exactness of the sequence-hence $\mathrm{H}_{2}(K)=0$.

For the case where $p=1$ we have:


Notice that $\mathbf{Z} \rightarrow \mathbf{Z} \oplus \mathbf{Z}$ maps $1 \mapsto(1,1)$ since it maps the generator of the circle $M \cap M^{\prime}$ to the generators of the disjoint circles $M$ and $M^{\prime}$. This mapping is injective, therefore
by exactness it follows that the map $\mathrm{H}_{1}(\mathrm{~K}) \rightarrow \mathbf{Z}$ is identically zero. From this we can extract the following short exact sequence:

therefore from the isomorphism theorem for groups we have

$$
\mathrm{H}_{1}(K) \simeq \frac{\mathbf{Z} \oplus \mathbf{Z}}{\operatorname{im}(\mathbf{Z} \rightarrow \mathbf{Z} \oplus \mathbf{Z})}=\frac{\mathbf{Z} \oplus \mathbf{Z}}{\{(2 x, 2 x): x \in \mathbf{Z}\}} \simeq \mathbf{Z} \oplus(\mathbf{Z} / 2 \mathbf{Z})
$$

Indeed, the last isomorphism is given by the mapping $[a, b] \mapsto(a+b,[b])$ : surjectivity is trivial, and on the other hand if $(a-b,[b])=(0,[0])$ then $a=b$ and $b=2 c$ for some $c \in \mathbf{Z}$ therefore $[a, b]=[2 c, 2 c]=0$, proving injectivity-which shows the last isomorphism. In summary we obtain:

$$
\mathrm{H}_{p}(K)= \begin{cases}\mathbf{Z}, & \text { if } p=0 \\ \mathbf{Z} \oplus(\mathbf{Z} / \mathbf{Z} \mathbf{Z}), & \text { if } p=1 \\ 0, & \text { otherwise }\end{cases}
$$

Example 19.7.9 (Projective plane). Consider $\mathbf{R P}^{2}$ as the gluing of the Möbius strip $M$ with the disk $D$ along the boundary. Let $p \in \boldsymbol{R} \mathbf{P}^{2}$ be any point, define $U$ to be a disk around $p$, and $V:=\mathbf{R} \mathbf{P}^{2} \backslash p \simeq_{\mathrm{h}} M \simeq_{\mathrm{h}} S^{1}$-then we also have $U \cap V=U \backslash p \simeq_{\mathrm{h}} S^{1}$. For any $p>2$ we have

and from exactness we obtain $\mathrm{H}_{p}\left(\mathbf{R} \mathbf{P}^{2}\right)=0$. For the case $p=2$ we have


From injectivity of the map $\delta$ we find $\mathrm{H}_{2}\left(\mathbf{R} \mathbf{P}^{2}\right) \simeq \operatorname{im} \delta$ and from exactness we have $\operatorname{im} \delta=\operatorname{ker}\left(i_{*} \oplus j_{*}\right)$. Notice however that the mapping $\mathbf{Z} \rightarrow \mathbf{Z}$ must send $1 \mapsto 2$ since the generator of $\mathrm{H}_{1}(U \cap V)$ is a loop through the circle, which is certainly a boundary in $\mathrm{H}_{1}(U)$ since $U$ is a disk, while $j$ will map such loop to a double loop around the boundary of the Möbius strip. This shows that $i_{*} \oplus j_{*}$ is injective and hence has null
kernel, proving that $\mathrm{H}_{2}\left(\mathbf{R P}^{2}\right)=0$. For the case where $p=1$ one has, in the reduced homology case, that:


Since $\mathbf{Z} \rightarrow \mathrm{H}_{1}\left(\mathbf{R P}^{2}\right)$ is surjective, there exists an isomorphism

$$
\mathrm{H}_{1}\left(\mathbf{R} \mathbf{P}^{2}\right) \simeq \mathbf{Z} / \operatorname{ker}\left(k_{*}-\ell_{*}\right)=\mathbf{Z} / \operatorname{im}\left(i_{*} \oplus j_{*}\right)=\mathbf{Z} / 2 \mathbf{Z}
$$

Moreover, since $\mathbf{R} \mathbf{P}^{2}$ is path-connected then $H_{0}\left(\mathbf{R} \mathbf{P}^{2}\right) \simeq \mathbf{Z}$. In summary we obtained:

$$
\mathrm{H}_{p}\left(\mathbf{R} \mathbf{P}^{2}\right)= \begin{cases}\mathbf{Z}, & \text { if } p=0 \\ \mathbf{Z} / 2 \mathbf{Z}, & \text { if } p=1 \\ 0, & \text { otherwise }\end{cases}
$$

## Part VI

## Homotopy Theory

## Chapter 20

## Simplicial Sets

### 20.1 Simplex Category

## Construction

Definition 20.1.1 (Skeletal simplex category). We denote by $\Delta$ the category whose objects are natural numbers $[n]:=\{0 \leqslant 1 \leqslant \ldots \leqslant n\}$, and whose morphisms $\phi:[n] \rightarrow$ [ $m$ ] are order-preserving-that is, if $i \leqslant j$ then $\phi(i) \leqslant \phi(j)$. The category $\Delta$ is called the skeletal simplex category.

Along with the simplex category comes distinguished morphisms:

- For each $0 \leqslant j \leqslant n$, we denote by $\delta_{j}^{n}:[n-1] \longmapsto[n]$ the order-preserving injective morphism skipping the $j$-th value:

$$
\delta_{j}^{n}(i):= \begin{cases}i, & \text { if } i<j, \\ i+1, & i \geqslant j .\end{cases}
$$

This morphism is called elementary faces.

- For each $0 \leqslant j \leqslant n$, we denote by $\sigma_{j}^{n}:[n+1] \rightarrow[n]$ the surjective morphism repeating the $j$-th value twice and every other value only once:

$$
\sigma_{j}^{n}(i):= \begin{cases}i, & \text { if } i \leqslant j, \\ i-1, & \text { if } i>j\end{cases}
$$

This morphism is called elementary degeneracies.
When convenient, we may drop the superscript of these maps and simply refer to them as $\delta_{i}$ and $\sigma_{j}$.

Notation 20.1.2 (Morphisms in $\Delta$ ). When convenient, a morphism $f:[n] \rightarrow[m]$ in $\Delta$ shall be explicitly denoted by

$$
f:=\left\langle f_{0} f_{1} \ldots f_{n}\right\rangle
$$

where $f_{j}:=f j \in[m]$ for each $0 \leqslant j \leqslant n$. Moreover these maps can also be referred to as simplicial operators.

Lemma 20.1.3 (Generating morphisms in $\Delta$ ). Every morphism of $\Delta$ can be generated by a composition of elementary faces and degeneracies.

Proof. Let $f:[n] \rightarrow[m]$ be any morphism of $\Delta$. We can factor $f$ through an injection $t:[n] \mapsto[k]$ and a surjection $s:[k] \rightarrow[m]:$

where $k:=|\operatorname{im} f|$. Since $s$ and $\iota$ must be order-preserving maps, it follows that they can be written as a finite composition of elementary degeneracies and elementary faces. $\quad$ b

Corollary 20.1.4. Every monomorphism of $\Delta$ is split (admits a retraction), while every epimorphism of $\Delta$ is split (admits a section).

Proof. Given an elementary face $\delta_{j}:[n] \mapsto[n+1]$ the elementary degeneracy map $\sigma_{j}:[n+1] \rightarrow[n]$ is a retract of $\delta_{j}$, and $\delta_{j}$ is a section of $\sigma_{j}$.

Corollary 20.1.5 (Cosimplicial identities). Fix any $n \in \mathbf{N}$ and consider indices $0 \leqslant i, j \leqslant$ $n$. The following identities correlate elementary faces and degeneracies:
(1) If $i<j$ then $\delta_{j}^{n} \delta_{i}^{n-1}=\delta_{i}^{n} \delta_{j-1}^{n-1}$.
(2) If $i<j$ then $\sigma_{i}^{n-1} \sigma_{j}^{n}=\sigma_{j-1}^{n-1} \sigma_{i}^{n}$.
(3) If $i<j$ then $\sigma_{i}^{n-1} \delta_{j}^{n}=\delta_{j-1}^{n-1} \sigma_{i}^{n-2}$.
(4) If $i=j-1$ or $i=j$, then $\sigma_{i}^{n} \delta_{j}^{n+1}=\operatorname{id}_{n}$.
(5) If $i>j$ then $\sigma_{i}^{n-1} \delta_{j}^{n}=\delta_{j}^{n-1} \sigma_{i-1}^{n-2}$.

These identities can be found in the following four commutative diagrams:


$$
[n] \underset{\sigma_{j-1}^{n-1}}{ }[n-1]
$$

$$
[n-1] \xrightarrow{\delta_{j}^{n}}[n]
$$

$$
\sigma_{i}^{n-2} \downarrow \quad \downarrow_{i}^{n-1}
$$

$$
[n-2] \underset{\delta_{j-1}^{n-1}}{\longrightarrow}[n-1]
$$

$$
\begin{gathered}
{[n-1] \xrightarrow[j]{\delta_{j}^{n}}[n]} \\
\left.\begin{array}{c}
\sigma_{i-1}^{n-2} \downarrow \\
{[n-2] \xrightarrow[\delta_{j}^{n-1}]{ }}
\end{array} \begin{array}{c} 
\\
\\
\\
\\
\sigma_{i}^{n-1}
\end{array}\right]
\end{gathered}
$$

Definition 20.1.6 (Cosimplicial object). Given a category C, we define a cosimplicial object in C to be a covariant functor

$$
F: \Delta \longrightarrow C
$$

Any cosimplicial object $F$ is completely determined by $F[n]$ for all $n \in \mathbf{N}$ and by the maps $F \delta_{i}$ and $F \sigma_{j}$. The collection of all cosimplicial objects of $C$ will be denoted by CoSimp(C).

## Limits \& Colimits in $\Delta$

Lemma 20.1.7 (Pushout of injections). Let $t:[k] \hookrightarrow[n]$ and $\tau:[k] \hookrightarrow[m]$ be orderpreserving inclusions where

$$
\iota(j):=j \text { and } \tau(j):=j+(m-k)
$$

that is, $\iota$ sends $[k]$ to the initial segment of $[n]$, while $\tau$ sends $[k]$ to the terminal segment of [ $m$ ]. There exists a pushout

in the simplex category $\Delta$.
Proof. Indeed, if we consider $[m+n-k]$ as a candidate for the pushout, notice that the following diagram commutes


Now, consider any object $[\ell] \in \Delta$ together with two morphisms $f:[n] \rightarrow[\ell]$ and $g:[m] \rightarrow[\ell]$ such that $f \iota=g \tau$. Define a map $\phi:[m+n-k] \rightarrow[\ell]$ as follows

$$
\phi(j):= \begin{cases}g(j), & \text { if } j \leqslant m-k \\ g(j)=f(j-(m-k)), & \text { if } m-k \leqslant j \leqslant m, \\ f(j-(m-k)), & \text { if } m \leqslant j \leqslant m+n-k\end{cases}
$$

It is easy to see that $\phi$ is an order-preserving map and is uniquely defined so that the following diagram commutes


Therefore $[m+n-k]=[n] \cup_{[k]}[m]$ since we are in a skeletal category and isomorphism classes contain a unique representative.

These pushouts lead to an interesting construction, any object $[n] \in \Delta$ is the colimit of a diagram consisting of [0]'s and [1]'s, since


Lemma 20.1.8 (Pushout of surjections). The following properties concern the pushout of pairs of surjective morphisms in $\Delta$ :
(a) Considering the cosimplicial identity (2) (see Corollary 20.1.5), where $i<j$, there exists sections $\alpha:[n] \rightarrow[n+1]$ of $\sigma_{i}^{n}$, and $\beta:[n-1] \rightarrow[n]$ of $\sigma_{i}^{n-1}$ such that the following diagram commutes

that is, $\sigma_{j}^{n} \alpha=\beta \sigma_{j-1}^{n-1}$ - these sections are said to be compatible with the square. Therefore, the square is an absolute pushout.
(b) Let $p:[n] \rightarrow[k]$ and $q:[n] \rightarrow[\ell]$ be surjections in $\Delta$. Then the pushout of $p$ and $q$ exists and is absolute:


Proof. For the proof of item (a), proceed as follows. From the cosimplicial identity (3) we have $\delta_{i}^{n+1} \sigma_{i}^{n}=\sigma_{i}^{n+1} \delta_{i+1}^{n+2}$, moreover, from (4) we obtain $\sigma_{i}^{n+1} \delta_{i+1}^{n+2}=\mathrm{id}_{n+1}$ - thus $\delta_{i}^{n+1}$ is a section of $\sigma_{i}^{n}$. Analogously, we find that $\delta_{i}^{n}$ is a section of $\sigma_{i}^{n-1}$. For the compatibility condition, notice that since $i<j$, then from (5) we know that $\sigma_{j}^{n} \delta_{i}^{n+1}=\delta_{i}^{n} \sigma_{j-1}^{n-1}$. Therefore we may define $\alpha:=\delta_{i}^{n+1}$ and $\beta:=\delta_{i}^{n}$.

For item (b), since surjections are finite compositions of degeneracies, via item (a) and Proposition 2.5.27, we conclude that the square in (b) is an absolute pushout. $\quad$,

Lemma 20.1.9. The following are properties of pullbacks along monomorphisms and pushouts along epimorphisms in the simplex category $\Delta$ :
(a) If $f:[m] \mapsto[n]$ is a monomorphism, then if $g:[k] \rightarrow[n]$ is any morphism such that $\operatorname{im} g \cap \operatorname{im} f \neq \varnothing$, then the square

is a pullback in $\Delta$.
(b) If $f:[m] \rightarrow[n]$ is an epimorphism, then for any morphism $g:[m] \rightarrow[k]$, then square

is a pushout in $\Delta$.

## I did not understand how to prove this, the proof presented on the book is a bit mysterious

### 20.2 Simplicial Sets

## Construction

Definition 20.2.1 (Simplicial object). For any category C, a simplicial object in C is a contravariant functor $\Delta^{\mathrm{op}} \rightarrow \mathrm{C}$. We define $\operatorname{Simp}(\mathrm{C})$ to be the category whose objects are simplicial objects and natural transformations between them-such category is commonly referred to as the simplicial category of $C$.

Notation 20.2.2 (Unwrapping the definition). Less compactly, a simplicial object $X \in$ Simp(C) consists of a collection of objects $\left(X_{n}\right)_{n \in \mathbf{N}}$, for $X_{n}:=X[n] \in C$, and arrows $\alpha^{*}: X_{n} \rightarrow X_{m}$ in C for each map $\alpha:[m] \rightarrow[n]$ in $\Delta$. The nature of these arrows are as discussed in Remark 1.7.3.

Notation 20.2.3 (Simplicial sets, notations and nomenclatures). In particular, the most important case we'll study for the time being is the simplicial category of sets, which we denote by sSet. The objects of sSet shall be called simplicial sets.

Given a simplicial set $X \in \operatorname{sSet}$, the points of $X_{n} \in$ Set will be referred to as the $n$-cells (or $n$-simplices) of $X \in$ sSet.

Since every map $\alpha$ in $\Delta$ may be decomposed into elementary face and degeneracies, the the collection $\delta_{i}^{*}$ and $\sigma_{j}^{*}$ also generate the morphisms $\alpha^{*}$ between the objects of
$\left(X_{n}\right)_{n \in \mathbf{N}}$ of C. In the context of the simplicial category, we denote them by

$$
\begin{aligned}
& d_{i}^{n}:=\left(\delta_{i}^{n}\right)^{*}: X_{n} \longrightarrow X_{n-1}, \\
& s_{j}^{n}:=\left(\sigma_{j}^{n}\right)^{*}: X_{n} \longrightarrow X_{n+1},
\end{aligned}
$$

for all $0 \leqslant i, j \leqslant n$. The arrows $d_{i}$ are called face maps (or cofaces), while the arrows $s_{j}$ are called degeneracy maps (or codegeneracies) of the simplicial object $X$.

A map $\eta: X \Rightarrow Y$ is a natural transformation between simplicial objects if and only if it is compatible with the face and degeneracy maps, that is, the following two diagrams commute in $C$ for all $n \in \mathbf{N}, 0 \leqslant i, j \leqslant n$ :


Corollary 20.2.4 (Simplicial identities). Since the face and degeneracy maps $d_{i}$ and $s_{j}$ are dual to the elementary face and degeneracy maps $\delta_{i}$ and $\sigma_{j}$, they satisfy the following identities-which are dual to Corollary 20.1.5:
(1) $d_{i} d_{j}=d_{j-1} d_{i}$, for $i<j$.
(2) $s_{j} s_{i}=s_{i} s_{j-1}$, for $i<j$.
(3) $d_{j} s_{i}=s_{i} d_{j-1}$, for $i<j-1$.
(4) $d_{j} s_{i}=\mathrm{id}$, if $i=j$ or $i=j-1$.
(5) $d_{j} s_{i}=s_{i-1} d_{j}$, for $i>j$.

## Subcomplexes

Definition 20.2.5 (Subcomplex). By a subcomplex of a simplicial set we mean a subobject.
Lemma 20.2.6. Given a morphism of simplicial sets $f: X \rightarrow Y$, the image $f X$ is a subcomplex of $Y$.

## Discrete Simplicial Sets

Definition 20.2.7 (Discrete simplicial set). We say that a simplicial set $X$ is discrete if every simplicial operator $f:[m] \rightarrow[n]$ induces a bijective set-function $f^{*}: X_{n} \rightarrow X_{m}$. We denote by sSet ${ }^{\text {disc }}$ the full subcategory of sSet consisting of discrete simplicial sets.

For every set $S$, there exists a discrete simplicial set $S^{\text {disc }}$ where $S_{n}^{\text {disc }}:=S$ for each $n \in \mathbf{N}$, and for any $f:[n] \rightarrow[m]$ we have $S^{\text {disc }} f:=\operatorname{id}_{s}$.

Corollary 20.2.8. For any set $S$ and simplicial set $X$, there exists a bijection

$$
\operatorname{Mor}_{\text {sSet }}\left(S^{\text {disc }}, X\right) \simeq \operatorname{Mor}_{\text {Set }}\left(S, X_{0}\right)
$$

Proof. Define a set-function $\Phi: \operatorname{Mor}_{\text {sSet }}\left(S^{\text {disc }}, X\right) \rightarrow \operatorname{Mor}_{\text {Set }}\left(S, X_{0}\right)$ by mapping each simplicial set morphism $f: S^{\text {disc }} \rightarrow X$ to the corresponding set function $f_{0}: S \rightarrow X_{0}$. To prove that $\Phi$ is injective, consider simplicial morphisms $f, g: S^{\text {disc }} \rightrightarrows X$ such that $f_{0}=g_{0}$. Applying the naturallity of $f$ and $g$ to the simplicial operator $j:[n] \rightarrow[0]$ we obtain that $j^{*} f_{0}=f_{n}$ and $j^{*} g_{0}=g_{n}$, but since $f_{0}=f_{0}$, then $f_{n}=g_{n}$-proving that $f=g$. Surjectivity is immediate, therefore $\Phi$ is a bijection.

Proposition 20.2.9 (sSet ${ }^{\text {disc } \& ~ S e t) . ~ T h e ~ f u l l ~ s u b c a t e g o r y ~ o f ~ d i s c r e t e ~ s i m p l i c i a l ~ s e t s ~ i s ~}$ equivalent to the category of sets.
Proof. We start by constructing a functor disc: Set $\rightarrow$ sSet ${ }^{\text {disc }}$ where we define the simplicial set $S^{\text {disc }}$ associated with a given set $S$ to be the simplicial set with $S_{n}^{\text {disc }}=S$ for every $S$. Each set-function $f: A \rightarrow B$ is mapped to their corresponding simplicial set morphism $f: A^{\text {disc }} \rightarrow B^{\text {disc }}$ given by $f_{n}:=f: A \rightarrow B$ for each $n \in \mathbf{N}$. We now show that disc is an equivalence of categories:

- (Fully faithful) Given any two sets, $S$ and $T$, define a set-function $\Phi$ : $\operatorname{Mor}_{\text {Set }}(S, T) \rightarrow$ $\operatorname{Mor}_{\text {sSet }}$ disc $\left(S^{\text {disc }}, T^{\text {disc }}\right)$ by mapping each set function $f: S \rightarrow T$ to $f: S^{\text {disc }} \rightarrow T^{\text {disc }}$ as above. Then it's clear that $\Phi$ is a bijection.
- (Essentially surjective) Given any $X \in$ sSet $^{\text {disc }}$, consider the set $S:=X_{0}$-we'll show that $X \simeq S^{\text {disc }}$. Define a morphism of simplicial sets $f: S^{\text {disc }} \rightarrow X$ where $f_{0}:=\operatorname{id}_{X_{0}}$. Let $\alpha:[n] \rightarrow[0]$ be any simplicial operator. Since $\alpha^{*}: X_{0} \xrightarrow{\simeq} X_{n}$ is a bijection, then by naturality of $f$ we find that

commutes in Set, making $f_{n}$ a bijection, which proves that $f$ is a natural isomorphism of simplicial sets.


## Standard Simplices

Definition 20.2.10 (Standard simplex). We define the standard $n$-simplex $\Delta^{n}$ to be the contravariant functor $\Delta^{\mathrm{op}} \rightarrow$ Set represented by $[n]$, that is, the simplicial set

$$
\Delta^{n}:=\operatorname{Mor}_{\Delta}(-,[n])
$$

The generator of the standard $n$-simplex corresponds to the $n$-cell

$$
\iota_{n}:=\operatorname{id}_{[n]} \in \Delta^{n}[n] .
$$

Let $\phi:[\ell] \rightarrow[k]$ be any simplicial operator, then the action of $\phi$ on the cells of $\Delta^{n}$ is given by the map $\Delta^{n} \phi=\phi^{*}: \Delta^{k} \rightarrow \Delta^{\ell}$. Explicitly, given a $k$-cell $\alpha:[k] \rightarrow[n]$ of $\Delta^{n}$, the action of the simplicial operator $\phi$ on $\alpha$ yields an $\ell$-cell

$$
\left(\phi^{*} \alpha=\alpha \phi:[\ell] \longrightarrow[n]\right) \in \Delta^{n}[\ell] .
$$

Lemma 20.2.11 (Yoneda consequences). Given a simplicial set $X$, there exists a bijection

$$
\operatorname{Mor}_{\text {sSet }}\left(\Delta^{n}, X\right) \simeq X_{n}
$$

which is explicitly given by the map $\eta \mapsto \eta_{n} \iota_{n}$. Equivalently, for each $n$-cell $x \in X_{n}$, there exists a unique morphism of simplicial sets $\eta^{x}: \Delta^{n} \rightarrow X$ such that $\eta_{n}^{x} \iota_{n}=x$.

Proof. Consider the Yoneda functor $H_{\Delta}: \Delta \rightarrow$ sSet, mapping each object $[n] \mapsto \Delta^{n}$ and each simplicial operator $f:[n] \rightarrow[m]$ to the morphism of simplicial sets $f_{*}: \Delta^{n} \rightarrow \Delta^{m}$. Explicitly, for each $k \in \mathbf{N}$ we have $\left(f_{*}\right)_{k}: \Delta^{n}[k] \rightarrow \Delta^{m}[k]$ mapping $\alpha \mapsto f \alpha$. By the Yoneda lemma we know that there exists, for each $n \in \mathbf{N}$, a bijection

$$
\operatorname{Mor}_{\mathrm{sSet}}\left(\Delta^{n}, X\right)=\operatorname{Mor}_{\mathrm{sSet}}\left(H_{\Delta}[n], X\right) \simeq X_{n}
$$

Definition 20.2.12 (Representing map of a cell). Regarding Lemma 20.2.11, we shall refer to $\eta^{x}$ as the representing map of the $n$-cell $x$. When convenient, we'll simply denote $\eta^{x}$ by $x: \Delta^{n} \rightarrow X$.

Corollary 20.2.13. There exists a bijection

$$
\operatorname{Mor}_{\mathrm{sSet}}\left(\Delta^{n}, \Delta^{m}\right) \simeq \Delta^{m}[n]
$$

The inverse of this set-function is explicitly given by mapping each simplicial operator $\alpha:[n] \rightarrow[m]$ to the corresponding morphism of simplicial sets $\alpha_{*}: \Delta^{n} \rightarrow \Delta^{m}$-where $\alpha_{*}$ acts on any $k$-cell $x \in \Delta^{n}[k]$ as $\alpha_{*} x=\alpha x \in \Delta^{m}[k]$.

Proof. In the proof of Lemma 20.2.11, merely consider the case where $X:=\Delta^{m}$. $\quad$.
Corollary 20.2.14 (Standard simplices \& $\Delta$ ). The full subcategory $\operatorname{sSet}_{\Delta}$ of sSet consisting of standard simplices is equivalent to the simplex category $\Delta$.

Proof. Consider the functor $F: \Delta \rightarrow \operatorname{sSet}_{\Delta}$ given by $[n] \mapsto \Delta^{n}$ and mapping each simplicial operator $f:[n] \rightarrow[m]$ to its corresponding morphism of simplicial sets $f_{*}: \Delta^{n} \rightarrow \Delta^{m}$. By Lemma 20.2.11 we see that $F$ is fully faithful and essentially surjective-thus an equivalence of categories.

Definition 20.2.15 (Standard simplices on totally ordered sets). Given a totally ordered set $S$, we define the standard $S$-simplicial set $\Delta^{S}$ to be the functor ${ }^{1}$

$$
\Delta^{S}:=\left.\operatorname{Mor}_{\mathrm{tOrd}}(-, S)\right|_{\Delta}
$$

that is, for each $n \in \mathbf{N}$ we define $\Delta_{n}^{S}$ to be the collection of order preserving maps $[n] \rightarrow S$.

[^29]Definition 20.2.16 (Boundary of a standard simplex). For every $n \in \mathbf{N}$, we define the boundary of the standard $n$-simplex $\Delta^{n}$ to be the subobject

$$
\partial \Delta^{n}:=\operatorname{colim}_{k \in[n]} \Delta^{[n] \backslash k}=\bigcup_{k \in[n]} \Delta^{[n] \backslash k}
$$

composed of all codimension-one faces of $\Delta^{n}$. Explicitly, for each $k \in \mathbf{N}$ we have

$$
\partial \Delta^{n}[k]=\left\{f \in \operatorname{Mor}_{\Delta}([k],[n]): f[k] \neq[n]\right\} \subseteq \Delta^{n}[k],
$$

composed of all non-surjective $k$-cells of $\Delta^{n}$.
Proposition 20.2.17. The boundary $\partial \Delta^{n}$ is the maximal proper subcomplex of $\Delta^{n}$.
Proof. Let $X$ be a proper subcomplex of $\Delta^{n}$.

## Prove

## Geometric Realization of the Simplex Category

We define, for each $n \in \mathbf{N}$, a corresponding standard topological $n$-simplex $\Delta_{\text {top }}^{n}$ given by

$$
\Delta_{\text {top }}^{n}:=\left\{\left(t_{0}, \ldots, t_{n}\right) \in \mathbf{R}^{n+1}: \sum_{j=0}^{n} t_{j}=1 \text { and } t_{j} \geqslant 0 \text { for all } j\right\}
$$

Each $\Delta_{\text {top }}^{n}$ is composed of $n+1$ vertices $v_{j}:=\left(\delta_{i j}\right)_{i=0}^{n}$.
From a categorical point of view, standard topological simplices are nothing more than a functor

$$
\Delta_{\text {top }}^{\bullet}: \Delta \longrightarrow \text { Top }
$$

mapping objects $\Delta_{\text {top }}^{\bullet}[n]:=\Delta_{\text {top }}^{n}$ —where $\Delta_{\text {top }}^{n}$ is endowed with the standard euclidean topology-and for each morphism $f:[n] \rightarrow[m]$ in $\Delta$, we map $\Delta_{\text {top }}^{\bullet} f:=f_{*}$, where $f_{*}: \Delta_{\text {top }}^{n} \rightarrow \Delta_{\text {top }}^{m}$ is a uniquely determined continuous map such that $f_{*}\left(v_{j}\right):=v_{f(j)}$. From this definition we obtain

$$
f_{*}\left(t_{0}, \ldots, t_{n}\right)=\left(s_{0}, \ldots, s_{m}\right), \text { where } s_{j}=\sum_{f(i)=j} t_{i}
$$

that is, $s_{j}$ is the sum of the points that are collapsed to the $j$-th coordinate.
This functor gives us a geometric visualization of the action of the elementary face and degeneracies:

- Given an elementary face map $\delta_{j}:[n-1] \mapsto[n]$, for any $0 \leqslant j \leqslant n$, the corresponding map $\left(\delta_{j}\right)_{*}: \Delta_{\text {top }}^{n-1} \hookrightarrow \Delta_{\text {top }}^{n}$ is given by

$$
\left(\delta_{j}\right)_{*} v_{i}= \begin{cases}v_{i}, & \text { if } i<j, \\ v_{i+1}, & \text { if } i \geqslant j .\end{cases}
$$

That is, $\left(\delta_{j}\right)_{*}$ embeds $\Delta_{\text {top }}^{n-1}$ as a face of $\Delta_{\text {top }}^{n}$ opposite to the $j$-th vertex.

- An elementary degeneracy map $\sigma_{j}:[n+1] \rightarrow[n]$, for any $0 \leqslant j \leqslant n$, has a map $\left(\sigma_{j}\right)_{*}: \Delta_{\text {top }}^{n+1} \rightarrow \Delta_{\text {top }}^{n}$ mapping the vertices as follows

$$
\left(\sigma_{j}\right)_{*} v_{i}= \begin{cases}v_{i}, & \text { if } i \leqslant j \\ v_{i-1}, & \text { if } i>j\end{cases}
$$

Geometrically, the degeneracy map makes $\Delta_{\text {top }}^{n+1}$ into $\Delta_{\text {top }}^{n}$ by removing a face of dimension 1 -through the projection parallel to the line connecting $v_{j}$ and $v_{j+1}$.

## Geometric Realization of a Simplicial Set

Given a simplicial set $X: \Delta^{\mathrm{op}} \rightarrow$ Set, we consider the topological space

$$
\coprod_{n \in \mathbf{N}} X_{n} \times \Delta_{\text {top }}^{n}
$$

and construct in this space a minimal equivalence relation $\sim_{g r}$ for which points $(x, t) \in$ $X_{n} \times \Delta_{\text {top }}^{n}$ and $\left(x^{\prime}, t^{\prime}\right) \in X_{m} \times \Delta_{\text {top }}^{m}$ are equivalent- $(x, t) \sim_{\text {gr }}\left(x^{\prime}, t^{\prime}\right)$-if and only if there exists a morphism $\alpha:[m] \rightarrow[n]$ in $\Delta$ such that $x^{\prime}=\alpha^{*} x$ and $t=\alpha_{*} t^{\prime}$. In an equivalent manner, we may summarize this equivalence relation as gluing points of the form

$$
\left(x, \alpha_{*} t\right) \sim_{\operatorname{gr}}\left(\alpha^{*} x, t\right) .
$$

The points of the resulting quotient space $\left(\coprod_{n \in \mathbf{N}} X_{n} \times \Delta_{\text {top }}^{n}\right) / \sim_{\text {gr }}$ are denoted by $x \otimes t-$ corresponding to the class of a pair $(x, t)$.

## Study why this construction is related to a tensor product of the form $X \otimes_{\Delta} \Delta_{\text {top }}^{\bullet}$.

Definition 20.2.18 (Geometric realization functor). We define the geometric realization of the category of simplicial sets to be a functor
mapping $|X|:=\left(\coprod_{n \in \mathbf{N}} X_{n} \times \Delta_{\text {top }}^{n}\right) / \sim_{\text {gr }}$ and for each natural transformation $\eta: X \rightarrow Y$ we have a topological morphism $|\eta|:|X| \rightarrow|Y|$ given by $x \otimes t \mapsto \eta_{n} x \otimes t$, for $(x, t) \in$ $X_{n} \times \Delta_{\text {top }}^{n}$.

For any simplicial set $X$, each $n$-cell $x \in X_{n}$ induces topological morphism

$$
\widehat{x}: \Delta_{\text {top }}^{n} \longrightarrow|X|, \text { mapping } t \longmapsto x \otimes t .
$$

From construction, given a morphism $\alpha:[n] \rightarrow[m]$ in $\Delta$ and a point $y \in X_{m}$ such that $y=\alpha^{*} x$, the diagram

commutes in Top.
This looks like there exists an induced slice over category $\mathrm{C} /|X|$-where C is a subcategory of Top consisting of standard topological simplices and morphisms between them-whose objects are $\widehat{x}$ and morphisms $\phi: \widehat{x} \rightarrow \widehat{y}$ are maps $\phi_{*}: \Delta_{\text {top }}^{m} \rightarrow \Delta_{\text {top }}^{n}$ for some $\phi:[n] \rightarrow[m]$ in $\Delta$.

Given any category $C$ and a functor $F: \Delta \rightarrow C$, we can define a simplicial set induced by any object $C \in C$ given by

$$
\operatorname{Mor}_{\mathrm{C}}(F(-), C): \Delta^{\mathrm{op}} \longrightarrow \text { Set. }
$$

This simplicial set maps each $[n] \in \Delta$ to the set of morphisms $\operatorname{Mor}_{C}(F n, C)$, and each morphism $\alpha:[m] \rightarrow[n]$ of $\Delta$ to the set-function $\alpha^{*}: \operatorname{Mor}_{C}(F n, C) \rightarrow \operatorname{Mor}_{C}(F m, C)$.

Definition 20.2.19 (Singular complex). Let $C$ be a category and $F: \Delta \rightarrow C$ be a functor. We define the singular complex of $C$ to be the functor

$$
\operatorname{Sing}_{F}: \mathrm{C} \longrightarrow \mathrm{sSet}
$$

mapping each object $C \in C$ to the simplicial set

$$
\operatorname{Sing}_{F}(C):=\operatorname{Mor}_{C}(F(-), C): \Delta^{\mathrm{op}} \longrightarrow \operatorname{Set},
$$

and each map $\alpha:[m] \rightarrow[n]$ to the set-function

$$
\operatorname{Sing}_{F}(\alpha):=\alpha^{*}: \operatorname{Mor}_{C}(F n, C) \longrightarrow \operatorname{Mor}_{C}(F m, C)
$$

In particular, the standard topological simplices functor $\Delta_{\text {top }}^{\bullet}: \Delta \rightarrow$ Top induces a singular complex on each topological space. Since we shall be mostly interested in this particular case for the time being, we shall reserve the notation

$$
\operatorname{Sing}:=\operatorname{Sing}_{\Delta_{\text {top }}}: \text { Top } \longrightarrow \text { sSet }
$$

with no subscripts, for the standard topological simplices functor. In this case, given a topological space $T$, we shall denote by $\operatorname{Sing}(T)_{n}$ the image of $[n] \in \Delta$ under the simplicial set Sing $(T)$.

Given a simplicial set $X$, the collection of simplexes $\left(\widehat{x}_{n}\right)_{n \in \mathbf{N}}$, where $x_{n} \in X_{n}$, covers the whole topological space $|X|$-in the sense that collection of images forms a cover of $|X|$. Therefore, given any topological morphism $\phi:|X| \rightarrow T$, this map is completely defined by the family of compositions $\left(\phi \widehat{x}_{n}: \Delta_{\text {top }}^{n} \rightarrow T\right)_{n \in \mathbf{N}}$. Given a morphism $\alpha:[n] \rightarrow[n]$ in $\Delta$, by Eq. (20.1), the diagram

commutes in Top. This construction induces unique a collection of maps

$$
\left(\phi_{n}: X_{n} \longrightarrow \operatorname{Sing}(T)_{n}\right)_{n \in \mathbf{N}},
$$

where $\phi_{n}(x) \mapsto \phi \widehat{x}$. Notice that this family of arrows is nothing more than a natural transformation $\phi: X \Rightarrow \operatorname{Sing}(T)$ between simplicial sets. From this we conclude that there exists a natural bijection

$$
\operatorname{Mor}_{T o p}(|X|, T) \simeq \operatorname{Mor}_{\mathrm{sSet}}(X, \operatorname{Sing}(T))
$$

thus the singular complex functor is right adjoint to the geometric realization,


## Geometric Realization as a CW-complex

Definition 20.2.20 (Degenerate $n$-cell). Given a simplicial set $X$, we say that an $n$-cell $x \in X_{n}$ is degenerate if $x \in s_{j} X_{n-1}$ for some codegeneracy map $s_{j}: X_{n-1} \rightarrow X_{n}$, where $0 \leqslant j \leqslant n-1$.

Equivalently, $x$ is denenerate if there exists a surjective map $\alpha:[n] \rightarrow[m]$ and $m$-cell $y \in X_{m}$ for which $x=\alpha^{*} y$.

Lemma 20.2.21 (Eilenberg-Zilber). Let $x$ be a $n$-cell of a given simplicial set $X$. There exists a unique pair $(\alpha, y)$ such that $\alpha:[n] \rightarrow[k]$ is a surjective map and $y$ is a nondegenerate $k$-cell of $X$ satisfying $\alpha^{*} y=x$.

Proof. The existence of the pair $(\alpha, y)$ comes straight from definition. Now suppose $(\beta, z)$ is another pair satisfying the said property, where $z$ is a non-degenerate $\ell$-cell of $X$. Since pushouts of a pair of surjections in $\Delta$ are absolute (see Lemma 20.1.8), the pushout of the pair $(\alpha, \beta)$ :

is turned into a pullback by the simplicial set $X: \Delta^{\mathrm{op}} \rightarrow$ Set, that is:


From the pushout property, $\gamma$ and $\omega$ are epimorphisms, thus split, $X$ ensures that $\gamma^{*}$ and $\omega^{*}$ are split-epimorphisms in Set. Therefore there exists s-cells $a, b \in X_{s}$ such that $\gamma^{*} a=y$ and $\omega^{*} b=z$. Notice however that we assumed $y$ and $z$ to be both nondegenerate, hence it must be the case that $\gamma$ and $\omega$ are identities. This implies in $\beta=\alpha$ and $y=z$.

Notice that, given a simplicial set $X$, we can naturally construct a filtration for $X$ (see Definition 17.2.2) by defining for each $n \in \mathbf{N}$ the subspace $|X|_{n}$ of $|X|$ given by

$$
|X|_{n}:=\left\{x \otimes t \in|X|:(x, t) \in X_{k} \times \Delta_{\text {top }}^{k} \text { for some } k \leqslant n\right\}
$$

so that the collection $\left(|X|_{n}\right)_{n \in \mathbf{N}}$ defines a filtration-that is, $|X|_{n} \leqslant|X|_{n+1}$ and $|X|=$ $\bigcup_{n \in \mathbf{N}}|X|_{n}$ is endowed with the weak topology (see Definition 17.1.21).

Lemma 20.2.22. Given a simplicial set $X$, the subspace $|X|_{0} \subseteq|X|$ is discrete and given by

$$
|X|_{0}=X_{0} \times \Delta_{\mathrm{top}}^{0}
$$

Proof. We shall create an isomorphism between both spaces. Define a surjective map $X_{0} \times \Delta_{\text {top }}^{0} \rightarrow|X|_{0}$ by mapping $(x, 1) \mapsto x \otimes 1$. Now, suppose that for some $x, y \in X_{0}$ one has $x \otimes 1=y \otimes 1$, then there must exist a pair $(z, t) \in X_{n} \times \Delta_{\text {top }}^{n}$, for some $n \in \mathbf{N}$, and parallel morphisms $\alpha, \beta:[0] \rightrightarrows[n]$ in $\Delta$ for which

$$
\left(x, \alpha_{*} 1\right)=\left(\alpha^{*} z, t\right) \quad \text { and } \quad\left(y, \beta_{*} 1\right)=\left(\beta^{*} z, t\right) .
$$

Since $\Delta_{\text {top }}^{0}=1$ is a single point, then $\alpha_{*}=\beta_{*}$ and hence $\alpha=\beta$, which shows that $x=y$.

Notation 20.2.23. Denote by $X_{n}^{\text {nd }}$ the collection of all non-degenerate $n$-cells of a given simplicial set $X$.

Lemma 20.2.24. Let $\xi \in|X|$ be any point, and let $n \in \mathbf{N}$ be the minimal index such that there exists $(x, t) \in X_{n} \times \Delta_{\text {top }}^{n}$ for which $\xi=x \otimes t$. Then $x$ is a non-degenerate simplex, and if $n \geqslant 1$ then $t \in \operatorname{Int} \Delta_{\text {top }}^{n}$-also, if that is the case, then the pair $(x, t)$ is unique with such property.

Proof. We shall prove this lemma in three steps:

- ( $x$ is non-degenerate) Suppose for the sake of contradiction that $x$ is degenerate, so that there exists a pair $(\alpha:[n] \rightarrow[m], y)$ with $y \in X_{m}$ and $m<n$ such that $\alpha^{*} y=x$. From this one obtains

$$
x \otimes t=\alpha^{*} y \otimes t=y \otimes \alpha_{*} t
$$

then $\left(y, \alpha_{*} t\right) \in X_{m} \times \Delta_{\text {top }}^{m}$ satisfies $\xi=y \otimes \alpha^{*} t$, which contradicts the assumption that $n$ was the minimal element of $\mathbf{N}$ with this property.

- $\left(t \in \operatorname{Int} \Delta_{\text {top }}^{n}\right)$ Assume that $n \geqslant 1$ and, for the sake of contradiction, suppose that $t \in \partial \Delta_{\text {top }}^{n}$-so that there exists at least one coordinate of $t$ that is zero. From this last comment, it follows that there exists an injective map $\beta:[k] \mapsto[n]$, for some $k<n$, such that $t \in \beta_{*} \Delta_{\text {top }}^{k}$. If $s \in \Delta_{\text {top }}^{k}$ is the point such that $\beta_{*} s=t$, then

$$
\xi=x \otimes t=x \otimes \beta_{*} s=\beta^{*} x \otimes s,
$$

but $\beta^{*} x \in \Delta_{\text {top }}^{k}$, which contradicts again the minimality of $n$.

- (Uniqueness of the pair $(x, t))$ Assume again that $n \geqslant 1$, and suppose that $\xi=$ $x \otimes t=y \otimes s$ for some $x, y \in X_{n}^{\text {nd }}$ and $t, s \in \operatorname{Int} \Delta_{\text {top }}^{n}$ —which can be assumed using the result of the last item. From the definition of $\otimes$ we may consider a finite zig-zag $\left(\alpha_{j}, \beta_{j}\right)_{j=1}^{N}$ in $\Delta$ :

where $n_{0}$ and $m_{N}$ are both defined to be $n$ and the intersection $\operatorname{im} \beta_{j} \cap \operatorname{im} \alpha_{j+1}$ is non-empty for every $1 \leqslant j<N$-together with a collection of pairs

$$
\left(\left(x_{j}, t_{j}\right),\left(y_{j}, s_{j}\right)\right)_{j=0}^{N} \in \prod_{j=0}^{N}\left(X_{k_{j}} \times \Delta_{\text {top }}^{k_{j}}\right) \times\left(X_{m_{j}} \times \Delta_{\text {top }}^{m_{j}}\right)
$$

for which one has the following relations:

$$
\left(x_{j},\left(\alpha_{j}\right)_{*} t_{j}\right)=\left(\alpha_{j}^{*} y_{j-1}, s_{j-1}\right) \quad \text { and } \quad\left(x_{j},\left(\beta_{j}\right)_{*} t_{j}\right)=\left(\beta_{j}^{*} y_{j}, s_{j}\right),
$$

and $\left(x_{0}, t_{0}\right):=(x, t)$, while $\left(y_{N}, s_{N}\right):=(y, s)$.
By hypothesis, since $t$ and $s$ are interior points of $\Delta_{\text {top }}^{n}$, then it must be the case that both $\alpha_{1}$ and $\beta_{N}$ are surjective, so that $k_{1}, k_{N} \geqslant n$. We proceed by induction on the zig-zag length $N$. If $N=1$, then the zig-zag is composed by two surjective maps $\alpha, \beta:[k] \rightrightarrows[n]$. Since pushouts of surjective maps exist in $\Delta$ and are absolute, if we consider the pushout of the pair $(\alpha, \beta)$ :

the contravariant functor $X$ maps this universal square to a pullback in Set:


Then since $\gamma$ is a split epimorphism, so is $\gamma^{*}$, which proves the existence of an $\ell$-cell $z \in X_{\ell}$ such that $\gamma^{*} z=x$. However, since $x$ is non-degenerate and $n$ is minimal, this can only happen if $\ell=n$ so that $\alpha, \beta=\operatorname{id}_{[n]}$-implying on $(x, t)=(y, s)$.

Assume, for the hypothesis of induction, that equality of the points is stablished for some $N-1>1$. We now work on the case $N>1$ : our goal will be to shorten the zig-zag sequence by one. Start by factoring the morphism $\beta_{1}$ as follows


If we now consider the pushout of the pair of surjective maps $\left(\alpha_{1}, \varepsilon\right)$, the simplicial set takes this pushout to a pullback in Set and from an analogous argument as done above, we conclude that we must have $\varepsilon=\operatorname{id}_{\left[k_{1}\right]}$. Therefore the factorization of $\beta_{1}$ reduces solely to $\beta_{1}=\delta$, which is a monomorphism. Since $\operatorname{im} \alpha_{2} \cap \operatorname{im} \beta_{1} \neq \varnothing$, then by Lemma 20.1.9 we know that the pullback of the pair $\left(\beta_{1}, \alpha_{2}\right)$ exists in $\Delta$ :

$$
\begin{gather*}
{\left[k_{1}^{\prime}\right] \stackrel{\theta}{\hookrightarrow}} \\
\eta \downarrow  \tag{20.2}\\
\downarrow \\
{\left[k_{1}\right]} \\
\stackrel{y}{\beta_{1}}
\end{gathered} \stackrel{\left[k_{2}\right]}{\longrightarrow} \begin{gathered}
\alpha_{2} \\
\left.m_{1}\right]
\end{gather*}
$$

Considering the simplex covariant functor $\Delta_{\text {top }}^{\bullet}$, we have a corresponding square

in Top, which is again a pullback.

## Prove that $\Delta_{\text {top }}^{\bullet}$ does indeed preserve the pullback

Therefore can find $c \in X_{k_{1}^{\prime}}$ such that $c=\eta^{*} x_{1}=\theta^{*} y_{1}$, and take the unique $w \in \Delta_{\text {top }}^{k_{1}^{\prime}}$ such that $\eta_{*} w=t_{1}$ and $\theta_{*} w=s_{1}$.
Notice that by the pullback Eq. (20.2) one has

so together with the pair $(c, w) \in X_{k_{1}^{\prime}} \times \Delta_{\text {top }}^{k_{1}^{\prime}}$ and $[n] \stackrel{\alpha_{1} \eta}{\longleftrightarrow}\left[k_{1}^{\prime}\right] \xrightarrow{\beta_{2} \theta}\left[m_{2}\right]$ we can shorten the zig-zag length by one, yielding the case $N-1$, which is true by hypothesis. Therefore $(x, t)=(y, s)$ and the lemma follows.

Theorem 20.2.25. Let $X$ be a simplicial set. The geometric realization of $X$ has a natural structure of a CW-complex with exactly one closed $n$-cell $\widehat{x}: \Delta_{\text {top }}^{n} \rightarrow|X|$ for each non-degenerate $n$-cell $x \in X_{n}$.

Proof. Let $(x, t),(y, s) \in \coprod_{x \in X_{n}^{\text {nd }}} \Delta_{\text {top }}^{n}$ be any two points. We analyse the case where $x \otimes t=y \otimes s$ are equivalent points of $|X|_{n}$ by means of Lemma 20.2.24:

- If both $t, s \in \operatorname{Int} \Delta_{\text {top }}^{n}$ then by the uniqueness of representatives we find that

$$
(x, t)=(y, s) .
$$

- Now if for instance $t \in \partial \Delta_{\text {top }}^{n}$, then there exits a unique $(z, r) \in X_{k} \times \operatorname{Int} \Delta_{\text {top }}^{k}$ for some minimal $k<n$ such that $x \otimes t=z \otimes t$. Notice however that since $k$ is minimal and unique with such property, then it must be the case that $s \in \partial \Delta_{\text {top }}^{n}$-therefore in this case one has both points

$$
(x, t),(y, s) \in \coprod_{x \in X_{n}^{\text {nd }}} \partial \Delta_{\text {top }}^{n}
$$

From this we can conclude that the following square is a pushout in Top:

which proves that $|X|$ is a $C W$-complex.
Definition 20.2.26 (Skeletal filtrations in sSet). Let $X$ be a simplicial set. We define the skeletal filtration of $X$ to be the collection $\left(\mathrm{sk}_{n} X\right)_{n \in \mathbf{N}}$ of simplicial sets, where $\mathrm{sk}_{n} X$ is the smallest subcomplex of $X$ containing every $k$-cell $x \in X_{k}$ for each $0 \leqslant k \leqslant n$. Explicitly, the collection of $k$-cells of $\mathrm{sk}_{n} X$ is

$$
\left(\operatorname{sk}_{n} X\right)_{k}=\bigcup_{0 \leqslant j \leqslant k}\left\{x f \in X_{k}: x \in X_{j} \text { and } f \in \operatorname{Mor}_{\Delta}([k],[j])\right\} .
$$

Therefore for each $n \in \mathbf{N}$ the simplicial set $\mathrm{sk}_{n} X$ is a subcomplex of $\mathrm{sk}_{n+1} X$.
From construction we see that

$$
X=\operatorname{colim}_{n \in \mathbf{N}} \operatorname{sk}_{n} X
$$

which is, loosely, the union of all $\mathrm{sk}_{n} X$.

Remark 20.2.27 (On Definition 20.2.26). Note by a subobject $S \in$ sSet of the simplicial set $X$ we actually mean an isomorphism class of natural monomorphisms-meaning, a class of natural transformations $\eta: S \Rightarrow X$ for which each associated morphism $\eta_{[n]}: S_{n} \mapsto X_{n}$ is a monomorphism in Set. We say that $(S, \eta)$ is equivalent to another subobject $(B, \xi: B \Rightarrow X)$ of $X$ if there exists a natural isomorphism $\sigma: S \xlongequal{\Longrightarrow} B$ such that, for any $[n] \in \Delta$, the diagram

commutes in Set.
Corollary 20.2.28 (Pushouts \& pullbacks in sSet). Let $X, Y$ and $Z$ be simplicial sets. We define the following:
(a) Given morphisms of simplicial sets $f: X \rightarrow Z$ and $g: Y \rightarrow Z$, the pullback of the pair $(f, g)$ to be the simplicial set $P$ defined as follows: for each $n \in \mathbf{N}$, the set $P_{n}$ is the pullback

in the category of sets.
(b) Given morphisms of simplicial sets $u: X \rightarrow Z$ and $v: Y \rightarrow Z$, the pushout of the pair $(u, v)$ to be the simplicial set $Q$ defined as follows: for each $n \in \mathbf{N}$, the set $Q_{n}$ is the pushout

in the category of sets.
Proof. These limits all exist in Set, therefore each limit is well defined in sSet.
Proposition 20.2.29. Given a simplicial set $X$, the square

is a pushout in sSet.

Proof. Let $P \in \mathrm{sSet}$ be the pushout of the given square, and define $p: P \rightarrow \mathrm{sk}_{n} X$ to be unique morphism of simplicial sets making the diagram

commute in sSet. We'll show that $p$ is an isomorphism of simplicial sets:

- (Epimorphism) Let $x \in\left(\operatorname{sk} X_{n}\right)_{k}$ be any $k$-cell, and let $(y, \alpha:[k] \rightarrow[m])$ be the unique representative pair of $x$ with $y \in X_{m}$ being a non-degenerate $m$-cell such that $x=y \alpha$ and $m \leqslant k$. If it is the case that $m<n$, then $x$ is a degenerate $k$-cell and also $y \in \operatorname{sk}_{n-1} X$-but then $x \in \operatorname{sk}_{n-1} X$ via $\alpha^{*}$, showing that $x \in \operatorname{im} p_{k}$. For the case where $m=n$ we see that $x$ is non-degenerate and therefore $\left(x, \Delta^{n}\right) \in$ $山_{x \in X_{n}^{\text {nd }}} \Delta^{n}$, thus $x \in \operatorname{im} p_{k}$ by the commutativity of

where $x: \Delta^{n} \rightarrow\left(\operatorname{sk}_{n} X\right)_{k}$ is the unique representative morphism of $x$.
- (Monomorphism) To prove that $p$ is a monomorphism it suffices to show the following two properties:
(i) Let $x \in X_{n}^{\text {nd }}$ be any non-degenerate $n$-cell and consider the canonical morphisms $x: \Delta^{n} \rightarrow \mathrm{sk}_{n} X$ and $\mathrm{sk}_{n-1} X \rightarrow \mathrm{sk}_{n} X$ —we'll show that the pullback of this pair of morphisms is $\partial \Delta^{n}$.
Let $\alpha \in \Delta^{n}[k]$ be a $k$-cell such that the $k$-cell $x \alpha \in X_{k}$ is an element of $\left(\mathrm{sk}_{n-1} X\right)_{k}$. Suppose, for the sake of contradiction, that $\alpha$ is an epimorphism, so that $n \leqslant k$. Since $x \alpha$ a $k$-cell of the $(n-1)$-skeleton, the unique pair $(y, \beta:[k] \rightarrow[m])$-where $y \in X_{m}$ is non-degenerate with $m<n$ and $y \beta=$ $x \alpha$-is such that $y \in\left(\mathrm{sk}_{n} X\right)_{m}$. Since $\alpha$ and $\beta$ are epimorphisms in $\Delta$, their pushout exists and is absolute:


Since $X$ is covariant, the corresponding square is a pullback in sSet:


Using the fact that $y \beta=x \alpha$, there exists $z \in X_{\ell}$ such that $z \delta=x$ and $z \gamma=y$. Since $x$ is assumed to be non-degenerate, and $\delta$ is an epimorphism with $\ell \leqslant n$ and $z \delta=x$, then it must be the case that $\delta=\mathrm{id}_{[n]}$ and hence $\ell=n$. Note however that $m<n$ so that $\gamma:[m] \rightarrow[n]$ should not be able to be an epimorphism, which is a contradiction. From this it follows that $\alpha$ isn't an epimorphism and hence $\alpha \in \partial \Delta^{n}[k]$ is a $k$-cell of the boundary. We thus conclude that

(ii) Given any two distinct non-degenerate $n$-cells $x, y \in X_{n}^{\text {nd }}$, let the square

be a pullback in sSet. Our goal will be to show that $v$ and $w$ can be factored through the monomorphism $\partial \Delta^{n} \mapsto \Delta^{n}$.
Consider a pair of $k$-cells $\alpha, \beta \in \Delta^{n}[k]$ such that $x \alpha=y \beta$. First, notice that if one of them was an identity map, then $k$ would coincide with $n$ and it would follow that both $\alpha$ and $\beta$ are identities, yielding $x=y$, which contradicts the assumption of $x$ and $y$ being distinct cells. We can therefore assume that neither $\alpha$ nor $\beta$ are identities. Consider the factorization


Since $\alpha_{-}$and $\beta_{-}$are epimorphisms, their pushout exists and is absolute:

$$
\begin{align*}
& {[k] \xrightarrow{\alpha_{-}}\left[m_{1}\right]} \\
& \beta_{-} \downarrow  \tag{20.3}\\
& \downarrow \\
& {\left[m_{2}\right] \xrightarrow[\gamma]{ } \quad \stackrel{\downarrow}{ }[\ell \delta}
\end{align*}
$$

Since $X$ is covariant, the corresponding square is a pullback in sSet:


Therefore there exists $z \in X_{\ell}$ for which $z \alpha_{-} \gamma=x$ while $z \delta \beta_{-}=y$. Now, since both $x$ and $y$ are non-degenerate cells, it follows that all maps in Eq. (20.3) are identity morphisms. It then follows that $\alpha=\alpha_{+}$and $\beta=\beta_{+}$ are monomorphisms. Since neither of them is an identity, it follows that $\alpha, \beta \in \partial \Delta^{n}$ are $k$-cells of the boundary. This shows that $v$ and $w$ can be both factorized through $\partial \Delta^{n} \mapsto \Delta^{n}$.

## Chapter 21

## Wheeled Graphs

### 21.1 Colouring the Portrait

Definition 21.1.1 (Colours \& Profiles). Fix, for the remaining of the chapter, a non-empty set $\mathfrak{C}$-whose elements will be called colours. A $\mathfrak{C}$-profile is a finite ${ }^{1}$ sequence of colours of $\mathfrak{C}$.

Notation 21.1.2. We denote a $\mathfrak{C}$-profile by $\underline{c}=\left(c_{1}, \ldots, c_{n}\right)$ or also $\mathcal{c}_{[1, n]}=\underline{c}$ when the indexing matters. Some of the operations on profiles are the following:

- The length of the profile is denoted by $|\underline{c}|=n$.
- Given $1 \leqslant j \leqslant n$, we define the notion of colour removal:

$$
\underline{c} \backslash c_{j}:=\left(c_{1}, \ldots, c_{i-1}, c_{i+1}, \ldots, c_{n}\right) .
$$

- Given another $\mathfrak{C}$-profile $\underline{d}$, with $|\underline{d}|=m$, we define the concatenation of $\underline{d}$ and $\underline{c}$ to be the $\mathfrak{C}$-profile

$$
(\underline{d}, \underline{c}):=\left(d_{1}, \ldots, d_{m}, c_{1}, \ldots, c_{n}\right) .
$$

- Given a permutation $\sigma \in \operatorname{Sym}_{n}$, we define the action of $\sigma$ on the profile $\underline{c}$ to be the $\mathfrak{C}$-profile $\sigma \underline{\mathcal{c}}$ given by

$$
\sigma \underline{c}:=\left(c_{\sigma(1)}, \ldots, c_{\sigma(n)}\right) .
$$

Definition 21.1.3 (C-profile category). We define a category $\operatorname{Prof}(\mathbb{C})$ whose objects are $\mathfrak{C}$-profiles, and a morphism $\sigma: \underline{c} \rightarrow \underline{d}$ is a permutation such that $\sigma \underline{\mathcal{c}}=\underline{d}$.

[^30]
## Chapter 22

## Model Categories

Definition 22.0.1 (Model category). Let $\mathcal{E}$ be a category. We say that $\mathcal{E}$ has a model structure if their morphisms are classified into three classes: fibrations (denoted by the arrow $\rightarrow$ ), cofibrations (denoted by $\rightarrow$ ), and weak equivalences (denoted by $\xrightarrow{\simeq}$ ), which satisfy the following properties:
(a) The category $\mathcal{E}$ is complete and cocomplete ${ }^{1}$.
(b) Given a commutative diagram

in $\mathcal{E}$, if any two of the maps $f, g$ and $h$ are weak equivalences, then the third is a weak equivalence.
(c) If $f$ is a retract-in the category $\operatorname{Mor}(\mathcal{E})$-of a morphism $g$, and $g$ is a weak equivalence, then $f$ is also a weak equivalence. For a shorter punchline, the classes of fibrations and cofibrations is closed under retracts.
(d) Given a commutative square

in $\mathcal{E}$ where $j$ is a cofibration and $p$ is a fibration, the dotted morphism $X \rightarrow E$ exists when either $j$ or $p$ is a weak equivalence. Said differently, fibrations have the right lifting property with respect to trivial cofibrations (both a cofibration and a weak equivalence), while trivial fibrations (both a fibration and a weak equivalence) have the right lifting property with respect to cofibrations.

[^31](e) Every morphism $f: X \rightarrow Y$ can be factored as

where $j$ is a cofibration and $q$ is a trivial fibration, and $i$ is a trivial cofibration and $p$ is a fibration.

The category $\mathcal{E}$ together with its model structure is said to be a model category.

## Part VII

## Manifold Theory

## Chapter 23

## Differentiable Manifolds

### 23.1 Differentiable Manifolds

## Charts \& Atlases

Definition 23.1.1 (Atlas). Let $X$ be an $n$-dimensional topological manifold. An atlas of class $C^{p}$ on $X$ is a collection of charts $\left\{\left(U_{j}, \phi_{j}\right)\right\}_{j \in J}$ such that
(a) For all $j \in J, U_{j} \subseteq X$ is an open set, and the collection $\left\{U_{j}\right\}_{j \in J}$ is an open cover for $X$.
(b) For every $j \in J, \phi_{j}: U_{j} \rightarrow V_{j}$ is a topological isomorphism from the open set $U_{j} \subseteq X$ to an open set $V_{j} \subseteq R^{n}$.
(c) For each pair $i, j \in J$, the induced change of coordinates map

$$
\phi_{j} \phi_{i}^{-1}: \phi_{i}\left(U_{i} \cap U_{j}\right) \stackrel{\simeq}{\leftrightharpoons} \phi_{j}\left(U_{i} \cap U_{j}\right)
$$

is of class $C^{p}$ — that is, every chart of the atlas with intersecting domain is compatible.
Definition 23.1.2 (Chart \& atlas compatibility). Let $X$ be a topological $n$-manifold. If $(U, \phi: U \xrightarrow{\simeq} V)$ and $\left(U^{\prime}, \psi: U^{\prime} \xrightarrow{\simeq} V^{\prime}\right)$ are charts on $X$, we say that they are $C^{p}$-compatible if the two induced transition maps

$$
\phi \psi^{-1}: \psi\left(U \cap U^{\prime}\right) \longrightarrow \phi\left(U \cap U^{\prime}\right) \quad \text { and } \quad \psi \phi^{-1}: \phi\left(U \cap U^{\prime}\right) \longrightarrow \psi\left(U \cap U^{\prime}\right)
$$

are of class $C^{p}$.
From this definition, we say that a chart is said to be compatible with a given atlas if it is compatible with every chart of the atlas. Moreover, given two atlases, we say that they are compatible if every chart of one is compatible with the other atlas.

Lemma 23.1.3. Let $\mathcal{A}:=\left\{\left(U_{j}, \phi_{j}\right)\right\}_{j \in J}$ be an atlas on a topological manifold $X$. If both $(V, \psi)$ and $(W, \sigma)$ are charts of $X$ compatible with the atlas $\mathcal{A}$, then they are compatible with each other.

Proof. Let $p \in V \cap W$ be any point and let $j \in J$ be such that $p \in U_{j}$ - thus $p \in V \cap W \cap U_{j}$. Since $\phi_{j} \psi^{-1}$ and $\sigma \phi_{j}^{-1}$ are $C^{p}$ maps, then $\sigma \psi=\left(\sigma \phi_{j}^{-1}\right) \circ\left(\phi_{j} \psi^{-1}\right)$ is $C^{p}$ when restricted to $\psi\left(V \cap W \cap U_{j}\right)$. Moreover, since $\psi(p) \in \psi\left(V \cap W \cap U_{j}\right)$, it follows that $\sigma \psi$ is $C^{p}$ on
$\psi(p)$ - therefore $\sigma \psi$ is $C^{p}$ on every point of its domain since $p$ was chosen arbitrarily. The same analogous argument can be used to show that $\psi \sigma^{-1}$ is $C^{p}$.

Proposition 23.1.4. The compatibility of atlases form an equivalence relation.
Proof. Clearly reflexivity and symmetry are satisfied. Let $\mathcal{U}:=\left\{\left(U_{j}, \phi_{j}\right)\right\}_{j \in J}$ and $\mathcal{V}:=\left\{\left(V_{i}, \psi_{i}\right)\right\}_{i \in I}$ be two compatible atlases for some topological manifold $X$. If $\mathcal{A}:=\left\{\left(A_{s}, \mu_{s}\right)\right\}_{s \in S}$ is another atlas for $X$, which happens to be compatible to $\mathcal{U}$, then for every $s \in S$ the maps $\phi_{j} \mu_{s}^{-1}$ and $\mu_{s} \phi_{j}^{-1}$ are of class $C^{p}$ for any $j \in J$. Since $\phi_{j} \psi_{i}^{-1}$ and $\psi_{i} \phi_{j}^{-1}$ are $C^{p}$ for all $i \in I$, then in particular

$$
\mu_{s} \psi_{i}^{-1}=\left(\mu_{s} \phi_{j}^{-1}\right) \circ\left(\phi_{j} \psi_{i}^{-1}\right) \quad \text { and } \quad \psi_{i} \mu_{s}^{-1}=\left(\psi_{i} \phi_{j}^{-1}\right) \circ\left(\phi_{j} \mu_{s}^{-1}\right)
$$

are both maps of class $C^{p}$. Therefore we conclude that $\mathcal{A}$ is compatible with $\mathcal{V}$. $\quad$.
Corollary 23.1.5. Any atlas on a topological manifold is contained in a unique maximal atlas - an atlas is said to be maximal if it isn't contained in any atlas other than itself.

## $C^{p}$-Manifolds

## Classical Definition

In this chapter we shall mostly consider the case of $C^{\infty}$-manifolds, also called smooth manifolds, but for generality we'll define differentiable manifolds for all $p \in \mathbf{N} \cup\{\infty\}$.

Definition 23.1.6 ( $C^{p}$-manifold structure on $X$ ). The equivalence classes of atlases of class $C^{p}$ on a topological space $X$ define what is called a $C^{p}$-manifold structure on $X$.

## Alternative Definition

We now give another definition of a $C^{p}$-manifold structure on topological spaces, to do that, we first introduce the following concept.

Definition 23.1.7 (Functionally structured space). Let $X$ be a topological space. A functional structure on $X$ is a map $F_{X}$ on the collection of open sets of $X$ such that, for any open set $U \subseteq X$, we have:
(a) $F_{X}(U)$ is a subalgebra of $C(U)$, the algebra of continuous real valued maps on $U$.
(b) $F_{X}(U)$ contains all constant maps.
(c) If $V \subseteq U$ is another open set of $X$, and $f \in F_{X}(U)$, then $\left.f\right|_{V} \in F_{X}(V)$.
(d) If $U=\bigcup_{j \in J} U_{j}$, and $f: U \rightarrow \mathbf{R}$ is a continuous map such that $\left.f\right|_{U_{j}} \in F_{X}\left(U_{j}\right)$ for each $j \in J$, then it follows that $f \in F_{X}(U)$.

The pair $\left(X, F_{X}\right)$ is called a functionally structured space.
Let $U \subseteq X$ be open. For any open set $V \subseteq U$ we define

$$
F_{U}(V):=F_{X}(V)
$$

and hence $\left(U, F_{U}\right)$ is a functionally structured space.

Definition 23.1.8 (Morphisms of functionally structured spaces). A morphism

$$
\phi:\left(X, F_{X}\right) \rightarrow\left(Y, F_{Y}\right)
$$

between functionally structured spaces is a map $\phi: X \rightarrow Y$ such that, for any open set $V \subseteq Y$ and $f \in F_{Y}(V)$, we have $f \phi \in F_{X}\left(\phi^{-1}(V)\right)$.

Definition 23.1.9 (Second definition of a $C^{p}$-manifold). An $n$-dimensional differentiable manifold is second countable functionally structured Hausdorff space ( $M, F$ ) which is locally isomorphic to $\left(\mathbf{R}^{n}, C^{p}\right)$.

The local isomorphism is equivalent to the requirement that, for each point $p \in M$, there exists a neighbourhood $U \subseteq M$ of $p$ such that $\left(U, F_{U}\right) \simeq\left(V, C_{V}^{p}\right)$ as functionally structured spaces - for some open set $V \subseteq \mathbf{R}^{n}$.

Lemma 23.1.10. Let $U, V \subseteq R^{n}$ be open subspaces. An isomorphism between functionally structured spaces

$$
\phi:\left(U, C_{U}^{p}\right) \xrightarrow{\simeq}\left(V, C_{V}^{p}\right)
$$

is a map $\phi: U \rightarrow V$ of class $C^{p}$ if and only if $f \phi \in C^{p}(U)$ for all $f \in C^{p}(V)$.
Proof. Clearly, if $\phi: U \rightarrow V$ is of class $C^{p}$ then $f \phi$ is a composition of $C^{p}$ maps, thus $f \phi \in C^{p}(U)$ for any $f \in C^{p}(V)$.

Conversely, if we have the hypothesis that $f \phi \in C^{p}(U)$ for all $f \in C^{p}(V)$, one may consider the canonical projections $\pi_{j}: \mathbf{R}^{n} \rightarrow \mathbf{R}$ and notice that $\pi_{j} \phi \in C^{p}(U)$. Therefore each component of $\phi$ is a $C^{p}(U)$ map, implying that $\phi$ itself is a $C^{p}(U, V)$ map. $\quad$.

Lemma 23.1.11 (Equivalence of the definitions). The constructions on Definition 23.1.6 and Definition 23.1.9 are equivalent.

Proof. Let $(M, F)$ be a $C^{p}$-manifold in the sense of Definition 23.1.9. A chart on $M$ will be interpreted as a pair $(U, \phi: U \xrightarrow{\simeq} V)$ such that $\phi$ is an isomorphism $(U, F) \simeq\left(V, C_{V}^{p}\right)$ of functionally structured spaces. Since every point of $M$ has a neighbourhood from which one can define the above mentioned isomorphism, we see that these charts do cover the whole space $M$.

It remains to be proven that the transition maps are $C^{p}$. From our interpretation of chart, given any two charts $\phi:\left(U, F_{U}\right) \xrightarrow{\leftrightharpoons}\left(V, C_{V}^{p}\right)$ and $\psi:\left(U^{\prime}, F_{U^{\prime}}\right) \xrightarrow{\simeq}\left(V^{\prime}, C_{V^{\prime}}^{p}\right)$ in $M$, since $\phi \psi^{-1}$ and $\psi \phi^{-1}$ are isomorphisms of functionally structured spaces $\left(V, C_{V}^{p}\right)$ and $\left(V^{\prime}, C_{V^{\prime}}^{p}\right)$ - by Lemma 23.1 .10 they are $C^{p}$ maps. Therefore, our collection of charts match the requirements of Definition 23.1.1.

For the converse, let $M$ be a $C^{p}$-manifold with an atlas $\mathcal{A}$. For every chart $(U, \phi: U \xrightarrow{\leftrightharpoons} V) \in \mathcal{A}$ (where $\phi$ is a topological morphism), define

$$
F(U):=\left\{f \phi \in C^{p}(U): f \in C^{p}(V)\right\}
$$

Let $x \in M$ be any point and consider a chart $(U, \psi: U \xrightarrow{\simeq} V) \in \mathcal{A}$, where $U \subseteq M$ is a neighbourhood of $x$. Notice that $\phi$ naturally induces a morphism of functionally structured spaces $\phi:\left(U, F_{U}\right) \rightarrow\left(V, C_{V}^{p}\right)$ - moreover, since $\phi$ is a topological isomorphism, then $\phi$ is an isomorphism $\left(U, F_{U}\right) \simeq\left(V, C_{V}^{p}\right)$.

Definition 23.1.12 (Orientation). A manifold $M$ is said to be oriented if for any two charts $\phi$ and $\psi$ on $M$, the Jacobian matrix of the transition map $\phi \psi^{-1}$ has a positive determinant for all points of its domain. This maximal atlas on $M$ is said to define an orientation on $M$. A manifold endowed with such an atlas is said to be orientable.

## Smooth Morphisms

## Morphisms Between Manifolds

Remark 23.1.13. From now on, unless stated otherwise, all manifolds are assumed to be endowed with a smooth structure.

Remark 23.1.14 (Chart notation). When convenient, a chart $(U, \phi: U \rightarrow V)$ shall simply be denoted by $\phi: U \rightarrow V$.

Definition 23.1.15 (Smooth map). Let $M$ and $N$ be smooth manifolds of dimension $m$ and $n$, respectively. A continuous map $f: M \rightarrow N$ is said to be smooth at a point $p \in M$ if there exists a chart $(V, \psi)$ about $f(p)$ in $N$, and a chart $(U, \phi)$ about $p$ in $M$ such that the map

$$
\psi f \phi^{-1}: \phi\left(f^{-1}(V) \cap U\right) \rightarrow \mathbf{R}^{n}
$$

where $\phi\left(f^{-1}(V) \cap U\right) \subseteq \mathbf{R}^{m}$, is smooth at $\phi(p)$. Naturally, $f$ is said to be smooth when $f$ is smooth at every point of $M$.

Remark 23.1.16. The requirement of continuity of $f: M \rightarrow N$ in Definition 23.1.15 is necessary to ensure that $f^{-1}(V) \subseteq M$ is open.

Lemma 23.1.17 (Smooth maps are choice-independent). Let $f: M \rightarrow N$ be a smooth map at $p \in M$ between smooth manifolds. If $(U, \phi)$ is any chart about $p$ in $M$ and $(V, \psi)$ is any chart about $f(p)$ in $N$, then the composition map $\psi f \phi^{-1}$ is $C^{\infty}$ at $\phi(p)$. Since both $M$ and $N$ are endowed with $C^{\infty}$ compatible charts, it follows that the transition $\operatorname{maps} \phi_{0} \phi^{-1}: \phi\left(U_{0} \cap U\right) \rightarrow \phi_{0}\left(U_{0} \cap U\right)$ and $\psi \psi_{0}^{-1}: \psi_{0}\left(V_{0} \cap V\right) \rightarrow \psi\left(V_{0} \cap V\right)$ are both $C^{\infty}$ maps. Therefore the composition

$$
\left(\psi \psi_{0}^{-1}\right) \circ\left(\psi_{0} f \phi_{0}^{-1}\right) \circ\left(\phi_{0} \phi^{-1}\right)=\psi f \phi^{-1}: \phi\left(U_{0} \cap U\right) \longrightarrow \mathbf{R}^{n}
$$

where $n$ is the dimension of $N$, is a $C^{\infty}$ map. Since $p \in U_{0} \cap U$, then $\psi f \phi^{-1}$ is $C^{\infty}$ at $\phi(p)$.

Proof. From the definition of smoothness, let $\left(U_{0}, \phi_{0}\right)$ be a chart about $p$ of $M$ and $\left(V_{0}, \psi_{0}\right)$ be a chart about $f(p)$ of $N$ such that the map $\psi_{0} f \phi_{0}^{-1}$ is $C^{\infty}$ at $\phi_{0}(p)$.

Proposition 23.1.18. The composition of smooth maps is smooth.
Proof. Let $N, M$ and $W$ be any three manifolds, and consider two smooth maps $f: N \rightarrow$ $M$ and $g: M \rightarrow W$. Take any three charts $(U, \phi)$ of $N,(V, \psi)$ of $M$, and $(E, \gamma)$ of $W$. Notice that the composition

$$
\left(\gamma g \psi^{-1}\right) \circ\left(\psi f \phi^{-1}\right)=\gamma(g f) \phi^{-1}: \phi\left(f^{-1}(V) \cap U\right) \longrightarrow \mathbf{R}^{w}
$$

where $w$ is the dimension of $W$, is a $C^{\infty}$ map since both $\gamma g \psi^{-1}$ and $\psi f \phi^{-1}$ are $C^{\infty}$. Since the charts where chosen arbitrarily, it follows that $g f$ is a smooth map. Notice that the last two conclusions came directly from Proposition 23.1.19.

## Equivalent Conditions for Smoothness

Proposition 23.1.19 (Equivalent conditions for smoothness). Let $f: N \rightarrow M$ be a continuous map between manifolds $N$ and $M$, with respective dimensions $n$ and $m$. The following properties are equivalent:
(a) The continuous map $f: N \rightarrow M$ is a smooth map.
(b) There exists atlases $\mathcal{N}$ for $N$ and $\mathcal{M}$ for $M$ such that, for every chart $(U, \phi) \in \mathcal{N}$ and $(V, \psi) \in \mathcal{M}$, the composition

$$
\psi f \phi^{-1}: \phi\left(f^{-1}(V) \cap U\right) \rightarrow \mathbf{R}^{m}
$$

is a $C^{\infty}$ map.
(c) For all pairs of charts $(U, \phi)$ on $N$ and $(V, \psi)$ on $M$, the composition

$$
\psi f \phi^{-1}:\left(f^{-1}(V) \cap U\right) \rightarrow \mathbf{R}^{m}
$$

is a $C^{\infty}$ map.
Proof. • (a) $\Rightarrow$ (c): Consider any pair of charts $(U, \phi)$ on $N$ and $(V, \psi)$ on $M$ for which $f^{-1}(V) \cap U$ is non-empty. Take any $p \in f^{-1}(V) \cap U$. Since $f$ is $C^{\infty}$, it follows that the composition $\psi f \phi^{-1}$ is $C^{\infty}$ at $\phi(p)$ - thus smooth for any point of its domain.

- $(c) \Rightarrow(b)$ : It suffices to choose atlases with $C^{\infty}$ compatible charts for $N$ and $M$.
- (b) $\Rightarrow$ (a): If $p \in N$ is any point, choose a chart $(U, \phi)$ about $p$ and $(V, \psi)$ about $f(p)$. Property (b) ensures that the composition $\psi f \phi^{-1}$ is, in particular, continuous at $\phi(p)$, since $p \in f^{-1}(V) \cap U$.

Corollary 23.1.20. Let $M$ be an $n$-manifold and $f: M \rightarrow \mathbf{R}^{d}$ be a continuous map. The following properties are equivalent:
(a) The map $f: M \rightarrow \mathbf{R}^{d}$ is a $C^{\infty}$-morphism.
(b) The manifold $M$ is endowed with an atlas such that, for every chart $\phi: U \rightarrow V$ of $M$, the map $f \phi^{-1}: \phi(U) \rightarrow \mathbf{R}^{d}$ is a map of class $C^{\infty}$, where $V \subseteq \mathbf{R}^{n}$.
(c) For all charts $\phi: U \rightarrow V$ of $M$, the map $f \phi^{-1}: \phi(U) \rightarrow \mathbf{R}^{d}$ is of class $C^{\infty}$, where $V \subseteq \mathbf{R}^{n}$.

Proof. The whole proposition is simply a particular case of Lemma 23.1.26.
Proposition 23.1.21 (Smoothness from projections). Let $M$ be an $n$-manifold. A map $f: M \rightarrow \mathbf{R}^{d}$ is a smooth map if and only if its projections $f_{j}: M \rightarrow \mathbf{R}$, for all $1 \leqslant j \leqslant d$, are smooth maps.

Proof. From definition, $f$ is a smooth map if and only if, for each chart $\phi: U \rightarrow V$ of $M$, the map $f \phi^{-1}: V \rightarrow \mathbf{R}^{d}$ is of class $C^{\infty}$. Moreover, from the definition of continuity on real spaces, we find that $f \phi^{-1}$ is $C^{\infty}$ if and only if $f_{j} \phi^{-1}$ is $C^{\infty}$ for all $1 \leqslant j \leqslant d$.

Proposition 23.1.22. Let $f: N \rightarrow M$ be a continuous map between two manifolds of dimensions $n$ and $m$, respectively. The following properties are equivalent:
(a) The map $f: N \rightarrow M$ is a $C^{\infty}$-morphism.
(b) The manifold $M$ is endowed with an atlas such that, for all charts $(V, \psi)$ of $M$, the $\operatorname{map} \psi f: f^{-1}(V) \rightarrow \mathbf{R}^{m}$ is a $C^{\infty}$-morphism.
(c) For every chart $(V, \psi)$ of $M$, the $\operatorname{map} \psi f: f^{-1}(V) \rightarrow \mathbf{R}^{m}$ is a $C^{\infty}$-morphism.

Proof. $(\mathrm{b}) \Rightarrow(\mathrm{a})$ : Together with the atlas of $M$, the continuous map $f$ induces a structure on $N$ as follows. For each chart $(V, \psi)$ of $M$, construct a collection $\mathcal{F}_{V}$ : for every chart $(U, \phi)$ in the atlas of $N$, define a chart $\left(U \cap f^{-1}(V),\left.\phi\right|_{U \cap f^{-1}(V)}\right)-$ define $\mathcal{F}_{V}$ to be the collection of all such chart. Then $\mathcal{F}_{V}$ is a smooth atlas for $f^{-1}(V) \subseteq N$. From (b) we know that $\psi f: f^{-1}(V) \rightarrow \mathbf{R}^{m}$ is of class $C^{\infty}$, then by Corollary 23.1.20 we have that the map $\psi f \phi^{-1}: \phi\left(U \cap F^{-1}(V)\right) \rightarrow \mathbf{R}^{m}$ is of class $C^{\infty}$. Now by Proposition 23.1.19 we obtain that $f$ is a smooth morphism.
(a) $\Rightarrow$ (c): Coordinate charts are $C^{\infty}$-morphisms and $f$ is a $C^{\infty}$-morphism by hypothesis. Therefore by Proposition 23.1.18 we find that $\psi f$ is a $C^{\infty}$-morphism. The implication (c) $\Rightarrow(\mathrm{b})$ is immediate.

Corollary 23.1.23. Let $f: N \rightarrow M$ be a continuous map between manifolds of dimensions $n$ and $m$, respectively. The following properties are equivalent:

1. The map $f: N \rightarrow M$ is a $C^{\infty}$-morphism.
2. The manifold $M$ is endowed with an atlas such that, for all charts $(V, \psi)$, the projections $\psi_{j} f: f^{-1}(V) \rightarrow \mathbf{R}$ of $f$ relative to the chart, for $1 \leqslant j \leqslant m$, are all $C^{\infty}$-morphisms.
3. For every chart $(V, \psi)$ of $M$, the components of $f$ with respect to the chart are smooth maps, that is, $\psi_{j} f: f^{-1}(V) \rightarrow \mathbf{R}$ are $C^{\infty}$-morphisms.

Proof. The proposition is consequence of Proposition 23.1.21 together with Proposition 23.1.22.

## The Category of Smooth Manifolds

Definition 23.1.24 (Category of $C^{\infty}$-manifolds). We define Man to be the category of smooth manifolds and smooth morphisms between them, these will be interchangeably called $C^{\infty}$-morphisms.

Corollary 23.1.25. An isomorphism in the category Man is a bijective $C^{\infty}$-morphism of manifolds with a smooth inverse. Some call these isomorphisms by "diffeomorphisms", we shall call them $C^{\infty}$-isomorphisms or smooth isomorphisms.

Lemma 23.1.26. Let $\left(M, F_{M}\right)$ and $\left(N, F_{N}\right)$ be smooth manifolds. A map $f: M \rightarrow N$ is smooth in the sense of Definition 23.1.15 if and only if $f$ is smooth in the sense of Definition 23.1.8.

Proof. Assume that $M$ and $N$ are, respectively, $m$ and $n$-dimensional spaces. First we consider $f$ as a morphism of functionally structured spaces. Let $p \in M$ be any point and consider charts, in the sense of isomorphisms of functionally structured spaces:

- $\phi: U \xrightarrow{\cong} V$ in $M —$ where $U$ is a neighbourhood of $p$ and $V \subseteq R^{m}$
- $\psi: U^{\prime} \xrightarrow{\cong} V^{\prime}$ in $N$ - where $U^{\prime}$ is a neighbourhood of $f(p)$ and $V^{\prime} \subseteq \mathbf{R}^{n}$.

Since $\phi$ is an isomorphism, one can consider its inverse and use the property that, for any open set $S \subseteq V$ and map $g \in F_{M}(S)$, then $g \phi^{-1} \in C^{\infty}(\phi(S)$ ). Since $f$ is a morphism of functionally structured spaces, in particular $\pi_{j} \psi \in F_{N}\left(U^{\prime}\right)$ for all $1 \leqslant j \leqslant n$, then $\pi_{j} \psi f \in F_{M}\left(f^{-1}\left(U^{\prime}\right)\right)$. Therefore, since $p \in U$ and $f(p) \in U^{\prime}$, the intersection $f^{-1}\left(U^{\prime}\right) \cap U \subseteq M$ is non-empty and we can conclude that

$$
\pi_{j} \psi f \phi^{-1} \in C^{\infty}\left(\phi\left(f^{-1}\left(U^{\prime}\right) \cap U\right)\right)
$$

Since this is the case for every $1 \leqslant j \leqslant n$, it follows that $\psi f \phi^{-1}$ is $C^{\infty}$ - therefore $f$ is smooth in the sense of Definition 23.1.15.

For the converse, suppose $f: M \rightarrow N$ is smooth as in Definition 23.1.15. From the last property of the functional structures on spaces, we can simply consider a chart $\psi: U^{\prime} \rightarrow V^{\prime}$ in $N$ and define a functional structure on $N$ as

$$
F_{N}\left(U^{\prime}\right):=\left\{g \phi \in C^{\infty}\left(U^{\prime}\right): g \in C^{\infty}\left(V^{\prime}\right)\right\} .
$$

If $h \in F_{N}\left(U^{\prime}\right)$ is any map, then the analogous structure $F_{M}$ on $M$ has

$$
F_{M}\left(f^{-1}\left(U^{\prime}\right)\right)=\left\{w f \in C^{\infty}\left(f^{-1}\left(U^{\prime}\right)\right): w \in C^{\infty}\left(U^{\prime}\right)\right\}
$$

therefore $h f \in F_{M}\left(f^{-1}\left(U^{\prime}\right)\right)$ as wanted.
Definition 23.1.27 (Lie group). A Lie group is a smooth manifold $G$ together with a multiplicative structure $G \times G \rightarrow G$ and an inverse $\operatorname{map} G \rightarrow G$, both of which are $C^{\infty}$-morphisms, making $G$ into a group.

## $C^{\infty}$-Isomorphisms

Proposition 23.1.28 (Coordinate maps are $C^{\infty}$-isomorphisms). Let $(U, \phi: U \rightarrow V)$ be a chart on an $n$-manifold $M$, then $\phi$ is a $C^{\infty}$-isomorphism.

Proof. By the definition, the coordinate map $\phi$ is a topological isomorphism, we thus show that both $\phi$ and $\phi^{-1}$ are smooth maps. Lets endow $U$ with the natural structure of a manifold via the smooth atlas $\{(U, \phi)\}$, now for $V$ we canonically endow it with the smooth atlas $\left\{\left(V, \mathrm{id}_{V}\right)\right\}$. For the smoothness of $\phi$, it suffices to observe that $\mathrm{id}_{V} \phi \phi^{-1}=\mathrm{id}_{V}$ is certainly $C^{\infty}$, therefore by Proposition 23.1.19 we find that $\phi$ is smooth. Analogously, for the smoothness of $\phi^{-1}: V \rightarrow U$, we know that $\phi \phi^{-1} \mathrm{id}_{V}=$ $\mathrm{id}_{V}$, thus $\phi^{-1}$ is smooth for the same reason.

Proposition 23.1.29 ( $C^{\infty}$-isomorphisms are charts). Let $M$ be an $n$-manifold and $U \subseteq M$ be any subset. If $f: U \rightarrow V$, where $V \subseteq \mathbf{R}^{n}$, is a $C^{\infty}$-isomorphism, the induced pair $(U, f)$ is a chart in the smooth structure of $M$.
Proof. Given any chart $(E, \phi)$ on $M$ with non-empty $E \cap U$, we have that $f \phi^{-1}$ and $\phi f$ are compositions of smooth maps (by Proposition 23.1.28 and Proposition 23.1.18), therefore both are smooth maps. This shows that $(U, f)$ is compatible with any intersecting chart of $M$, therefore by the maximality of the smooth atlas of $M$ we find that $(U, f)$ is a chart of $M$.

Proposition 23.1.30. Let $U \subseteq M$ be an open set in the $n$-dimensional smooth manifold $M$. If $F: U \rightarrow B$ is a $C^{\infty}$-isomorphism where $B \subseteq \mathbf{R}^{n}$ is an open set, then the pair $(U, F)$ is a coordinate chart in the differentiable structure of $M$.

Proof. Given any chart $(V, \phi)$ of the manifold $M$ where $U \cap V$ is non-empty, one has that the transition maps $F \phi^{-1}$ and $\phi F^{-1}$ are both compostions of $C^{\infty}$-isomorphisms-thus themselves $C^{\infty}$-isomorphisms. Since the differentiable structure of $M$ is defined to be maximal, it must be the case that $(U, F)$ is a chart in the atlas of $M$.

## Local Coordinates

Let $M$ be an smooth $n$-manifold and $f: M \rightarrow \mathbf{R}$ be any real valued map. If $x: U \rightarrow V$ is a chart of $M$, we consider the induced map $\bar{f}:=f x^{-1}: V \rightarrow \mathbf{R}$, a multivariable real map from a subset of $\mathbf{R}^{n}$ to $\mathbf{R}$. For any $p \in U$, one has "coordinates" $x(p)=\left(x_{1}(p), \ldots, x_{n}(p)\right)$. These coordinates can be used to compute $f$ via

$$
f(p)=\bar{f}\left(x_{1}(p), \ldots, x_{n}(p)\right)
$$

working as some kind of local coordinates on the open set $U$. This an intuitive view of what the following definition states.

Definition 23.1.31 (Local coordinates). If $X$ is an $\mathbf{R}^{n}$ modelled manifold, and $\phi: U \rightarrow \mathbf{R}^{n}$ is a chart, where $U \subseteq X$ is open, then, we say that the collection $\left(\phi_{j}\right)_{j=1}^{n}$ are local coordinates for $X$ on $U$.

## Partial Derivatives

Definition 23.1.32 (Partial derivative). Let $M$ be an $n$-manifold, and $(U, \phi)$ be a chart of $M$. Given a $C^{\infty}$-morphism $f: M \rightarrow \mathbf{R}$ we define, for all $p \in U$, the partial derivative of $f$ with respect to $\phi_{j}$ at $p$ to be

$$
\left.\frac{\partial}{\partial \phi_{j}}\right|_{p} f:=\frac{\partial f}{\partial \phi_{j}}(p):=\frac{\partial\left(f \phi^{-1}\right)}{\partial \pi_{j}}(\phi(p)):=\left.\frac{\partial}{\partial \pi_{j}}\right|_{\phi(p)}\left(f \phi^{-1}\right) .
$$

where $\phi_{j}$ is the $j$-th projection of $\phi$, and $\pi_{j}: \mathbf{R}^{n} \rightarrow \mathbf{R}$ is the $j$-th canonical projection, for $1 \leqslant j \leqslant n$. Since $\phi$ is a bijective map, we have that

$$
\frac{\partial f}{\partial \phi_{j}} \circ \phi^{-1}=\frac{\partial\left(f \phi^{-1}\right)}{\partial \pi_{j}}: \phi(U) \longrightarrow \mathbf{R}
$$

is a map of class $C^{\infty}$ on $\phi(U)$. Moreover, since its pullback is $C^{\infty}$, then $\partial f / \partial \phi_{j}$ is a $C^{\infty}$-morphism on $U$.

Proposition 23.1.33. Let $M$ be an $n$-manifold and $(U, \phi)$ be a chart of $M$. Then we have that for all $1 \leqslant i, j \leqslant n$, the projections of $\phi$ satisfy

$$
\frac{\partial \phi_{i}}{\partial \phi_{j}}=\delta_{i j}
$$

Proof. From the definition, one has

$$
\frac{\partial \phi_{i}}{\partial \phi_{j}}(p)=\frac{\partial\left(\phi_{i} \phi^{-1}\right)}{\partial \pi_{j}}(\phi(p))=\frac{\partial\left(\left(\pi_{i} \phi\right) \circ \phi^{-1}\right)}{\partial \pi_{j}}(\phi(p))=\frac{\partial \pi_{i}}{\partial \pi_{j}}(\phi(p))=\delta_{i j}
$$

for any $p \in U$.
Definition 23.1.34 (Jacobian). Let $f: N \rightarrow M$ be a $C^{\infty}$-morphism - assume $n$ and $m$ are the respective dimensions of $N$ and $M$ - and consider a pair of charts $(U, \phi)$ of $N$ and $(V, \psi)$ of $M$, for which $f(U) \subseteq V$. We define the $i$-th projection $f$ in the chart $(V, \psi)$ to be the map

$$
f_{i}:=\psi_{i} f: U \longrightarrow \mathbf{R},
$$

for all $1 \leqslant i \leqslant m$. We also define the Jacobian matrix of $f$ in the chart $(V, \psi)$ to be the $m \times n$ matrix whose $(i, j)$-th component is $\partial f_{i} / \partial \phi_{j}$ - for $1 \leqslant i \leqslant m$ and $1 \leqslant j \leqslant n$.

If it is the case that $N$ and $M$ have the same dimension, we define the Jacobian of $f$ in the chart $(V, \psi)$ to be the determinant of its respective Jacobian matrix.

## Inverse Map Theorem for Manifolds

Theorem 23.1.35 (Inverse map theorem). Let $f: N \rightarrow M$ be a $C^{\infty}$-morphism, $p \in N$ be any point, and both $N$ and $M$ be $n$-dimensional manifolds. Let $(U, \phi)$ be a chart about $p$ in $N$, and $(V, \psi)$ be a chart about $f(p)$ in $M$, such that $f(U) \subseteq V$. Then $f$ is locally invertible at $p$ if and only if the determinant $\operatorname{det}\left[\frac{\partial f_{i}}{\partial \phi_{j}}\right]_{i j}$ is non-zero-where $f_{i}:=\psi_{i} f$.

Proof. Let $\pi_{j}: \mathbf{R}^{n} \rightarrow \mathbf{R}$ denote the $j$-th canonical projection, so that $f_{i}=\psi_{i} f=\pi_{i} \psi f$. The local representation of $f_{* p}$ relative to the charts $(U, \phi)$ and $(V, \psi)$ is given by

$$
\begin{equation*}
\left[\frac{\partial f_{i}}{\partial \phi_{j}}\right]_{i, j=1}^{n}=\left[\frac{\partial\left(\pi_{i} \psi f\right)}{\partial \phi_{j}}\right]_{i, j=1}^{n}=\left[\frac{\partial\left(\pi_{i} \psi f \phi^{-1}\right)}{\partial \pi_{j}}(\phi(p))\right]_{i, j=1}^{n} \tag{23.1}
\end{equation*}
$$

That is, the local representation of $f_{* p}$ is the same as the Jacobian matrix of

$$
\psi f \phi^{-1}: \phi\left(f^{-1} V \cap U\right) \rightarrow \mathbf{R}^{m}
$$

at the point $\phi(p)$. By the inverse map theorem for Banach spaces (see Theorem A.5.8), we find that the determinant of the Jacobian Eq. (23.1) is non-zero if and only if $\psi f \phi^{-1}$ is a locally invertible map at $\phi(p)$. Notice however that both $\phi$ and $\psi$ are $C^{\infty}$-isomorphisms, thus $\psi f \phi^{-1}$ is locally invertible at the said point if and only if $f$ is invertible at $p$.

Corollary 23.1.36. Let $M$ be an $n$-dimensional manifold, and $(U, \phi)$ be a chart of $M$ about a given point $p \in M$. A $C^{\infty}$-morphism $F: U \rightarrow \mathbf{R}^{n}$, in the coordinate chart $(U, \phi)$, forms a coordinate chart ${ }^{1}$ about $p$ if and only if its Jacobian determinant $\operatorname{det}\left[\frac{\partial F_{i}}{\partial \phi_{j}}\right]_{i j}$ is non-zero.

Proof. From the inverse map theorem, the determinant of the Jacobian of $F$ at the point $p$ is non-zero if and only if $F: U \rightarrow \mathbf{R}^{n}$ is locally invertible at $p$. Moreover, the condition for $F$ to be locally invertible is equivalent to the existence of a neighbourhood $X \subseteq M$ of $p$ such that the induced map $F: X \rightarrow F(X)$ is a $C^{\infty}$-isomorphism, and thus $(X, F)$ is a coordinate chart about $p$ (see Proposition 23.1.30).

### 23.2 Structures

## Classical Approach

## Functional Structured Approach

Definition 23.2.1 (Induced structure). Let $\left(X, F_{X}\right)$ be a functionally structured space, and $\phi: X \rightarrow Y$ be a continuous map, where $Y$ is a topological space. We define the induced functional structure on $Y$ via $F_{X}$ and $\phi$ to be given by

$$
F_{Y}(U):=\left\{f \in C(U): f \phi \in F_{X}\left(\phi^{-1}(U)\right)\right\}
$$

for any open set $U \subseteq Y$.
Definition 23.2.2 (Induced structure on subspace). Let $(X, F)$ be a functionally structured space, and $A \subseteq X$ be a subspace. We construct a functional structure $F_{A}$ on $A$ as follows: for all open sets $U \subseteq A$, a continuous map $f: U \rightarrow \mathbf{R}$ is contained in $F_{A}(U)$ if and only if for every $p \in U$ there exists a neighbourhood $W \subseteq X$ of $p$ such that $f$ is the restriction to $W \cap A$ of some map $g \in F(W)$.

Definition 23.2.3 (Binary product of manifolds). Let $M$ and $N$ be manifolds of dimension $m$ and $n$, respectively. If $\phi: U \rightarrow \mathbf{R}^{m}$ is a chart for $M$ and $\psi: V \rightarrow \mathbf{R}^{n}$ is a chart for $N$, we take $\phi \times \psi: U \times V \rightarrow \mathbf{R}^{m+n}$ to be a chart for the product space $M \times N$. The maximal atlas consisting of such product charts make $M \times N$ into a product manifold of dimension $m+n$.

Definition 23.2.4 (Submanifold). Let $M$ be an $n$-manifold. A subset $N \subseteq M$ is said to be a $k$-manifold of $M$ if for each $p \in N$ there exists a chart $\phi: U \rightarrow V$ of $M$ about $p$ for which

$$
\phi(U \cap N)=V \cap\left(\mathbf{R}^{k} \times\{0\}\right)
$$

Charts with this property are called adapted to $N$. The maximal atlas containing all adapted charts to $N$ makes $N$ into a manifold.

[^32]We define the codimension of $N$ in $M$ to be the difference $n-k$. The submanifold $N$ is said to be smooth exactly when about each point there exists an adapted smooth chart to $N$ - so that one can find a unique maximal smooth atlas for $N$.

Definition 23.2.5 (Smooth embedding). A $C^{\infty}$-morphism $f: N \rightarrow M$ is said to be a smooth embedding if the image $f(N) \subseteq M$ is a smooth submanifold of $M$ and the induced restriction $f: N \xrightarrow{\leftrightharpoons} f(N)$ is a $C^{\infty}$-isomorphism.

Example 23.2.6 (Sphere). Consider the sphere $S^{n} \subseteq \mathbf{R}^{n+1}$, we'll define a smooth atlas on $S^{n}$ using solely two charts. To that end, define two points $n:=(0, \ldots, 0,1)$ and $s:=(0, \ldots, 0,-1)$ in $\mathbf{R}^{n+1}$ - these correspond to the north and south poles of $S^{n}$, respectively. Consider the stereographic projection $\phi_{n}: S^{n} \backslash\{n\} \rightarrow \mathbf{R}^{n}$, from the sphere with its north pole cut out to the hyperplane $R^{n} \simeq \mathbf{R}^{n} \times\{0\}$ - the stereographic projection maps each point $x \in S^{n}$ to the intersection of the line passing through $x$ and $n$, and the hyperplane above-mentioned. The inverse map of $\phi_{n}$ is $\pi_{n}: \mathbf{R}^{n} \rightarrow S^{n} \backslash\{n\}$, mapping

$$
x \longmapsto \frac{\left(2 x,\|x\|^{2}-1\right)}{(1+\|x\|)^{2}}
$$

We define the south stereographic projection $\phi_{s}: S^{n} \backslash\{s\} \rightarrow \mathbf{R}^{n}$ equivalently. This construction has a smooth transition map

$$
\phi_{s} \phi_{n}^{-1}(x)=\frac{y}{\|y\|^{2}}
$$

### 23.3 Partitions of Unity

Definition 23.3.1 (Adequate atlas). Given an $m$-manifold $M$, we say that an atlas $\left(U_{j}, \phi_{j}\right)_{j \in J}$ is adequate if it is locally finite, and for all we have either $\phi_{j}\left(U_{j}\right)=\mathbf{R}^{m}$ or $\phi_{j}\left(U_{j}\right)=\mathbf{H}^{m}$, and

$$
\bigcup_{j \in J} \phi_{j}^{-1}\left(\operatorname{Int} D^{m}\right)=M
$$

Since every topological manifold is paracompact, by Corollary 14.5 .8 we obtain the following corollary.

Corollary 23.3.2. Let $\mathcal{U}$ be a covering of $M$. There exists an adequate atlas $\left(U_{j}, \phi_{j}\right)_{j \in J}$ such that $\left(U_{j}\right)_{j \in J}$ is a refinement of $\mathcal{U}$.

Definition 23.3.3 (Bump functions \& partition of unity). Let $M$ be an $m$-manifold and $\left(U_{j}, \phi_{j}\right)_{j \in J}$ be an adequate atlas on $M$. Let $\lambda: \mathbf{R}^{m} \rightarrow \mathbf{R}$ be a smooth non-negative map such that $\left.\lambda\right|_{D^{m}}:=1$ and $\left.\lambda\right|_{\mathbf{R}^{m} \backslash D^{m}(2)}:=0$.

- Define, for every $j \in J$, maps $\lambda_{j}: M \rightarrow \mathbf{R}$ where

$$
\left.\lambda_{j}\right|_{U_{j}}:=\lambda \phi_{j} \quad \text { and }\left.\quad \lambda_{j}\right|_{M \backslash U_{j}}:=0
$$

We name the collection $\left(\lambda_{j}\right)_{j \in J}$ the bump functions associated to $\left(U_{j}\right)_{j \in J}$.

- Define, for each $j \in J$, maps $\mu_{j}: M \rightarrow \mathbf{R}$ given by

$$
\mu_{j}(p):=\frac{\lambda_{j}(p)}{\sum_{i \in J} \lambda_{i}(p)}
$$

for all $p \in M$. The collection $\left(\mu_{j}\right)_{j \in J}$ is said to be the partition of unity associated to the atlas $\left(U_{j}, \phi_{j}\right)_{j \in J}$.

### 23.4 Tangent Space

## Algebra of Germs

Consider any point $p \in M$. Define an equivalence relation $\sim_{p}$ on the set of realvalued $C^{\infty}$-morphisms in $M$ as follows: given neighbourhoods $U$ and $V$ of $p$, and $C^{\infty}$-morphisms $f: U \rightarrow \mathbf{R}$ and $g: V \rightarrow \mathbf{R}$, we say that $f \sim_{p} g$ if and only if there exists a neighbourhood $Q \subseteq U \cap V$ of $p$ such that $\left.f\right|_{Q}=\left.g\right|_{Q}$.

Definition 23.4.1 (Algebra of germs in a manifold). Given a manifold $M$ and a point $p \in M$, we define the germs of real-valued $C^{\infty}$-morphisms at $p$, to be the natural R-algebra on the following quotient:

$$
C_{p}^{\infty}(M):=\left\{f \in \operatorname{Mor}_{\operatorname{Man}}(U, \mathbf{R}): p \in U \subseteq M\right\} / \sim_{p}
$$

A point-derivation of $C_{p}^{\infty}(M)$ is an $\mathbf{R}$-linear map $X: C_{p}^{\infty}(M) \rightarrow \mathbf{R}$ such that

$$
X(f g)=(X f) g(p)+f(p)(X g)
$$

Definition 23.4.2 (Tangent vector \& tangent space). Let $M$ be a manifold and $p \in M$ be any point. A tangent vector at $p$ is a point-derivation on $C_{p}^{\infty}(M)$. The collection of tangent vectors at $p$, denoted by $T_{p} M$, together with a natural action $\mathbf{R} \times T_{p} M \rightarrow T_{p} M$, forms the $\mathbf{R}$-vector space called tangent space of $M$ at $p$.

## Differential of a Smooth Morphism

Definition 23.4.3 (Differential at a point). Let $f: N \rightarrow M$ be a $C^{\infty}$-morphism. At every point $p \in N$, the morphism $f$ induces a pushforward

$$
f_{*}: T_{p} N \longrightarrow T_{f(p)} M,
$$

called differential of $f$ at $p$, which is an $\mathbf{R}$-linear map between tangent spaces. For any tangent vector $X_{p}: C_{p}^{\infty}(N) \rightarrow \mathbf{R}$ of $T_{p} N$, we define the induced tangent vector $f_{*}\left(X_{p}\right): C_{f(p)}^{\infty}(M) \rightarrow \mathbf{R}$ of $T_{f(p)} M$ as the $\mathbf{R}$-linear map given by

$$
\left(f_{*}\left(X_{p}\right)\right)(g):=X_{p}(g f) \in \mathbf{R},
$$

for all $g \in C_{f(p)}^{\infty}(M)$.

Proposition 23.4.4. Let $f: N \rightarrow M$ and $g: M \rightarrow P$ be $C^{\infty}$-morphisms of manifolds, and $p \in N$ be a given point. Then one has that the differential of the composition $g f$ at $p$, the linear map $(g f)_{*}: T_{p} N \rightarrow T_{f(p)} M$, satisfies

$$
(g f)_{* p}=g_{* f(p)} f_{* p}
$$

where $g_{* f(p)}$ denotes the differential of $g$ at $f(p)$, while $f_{* p}$ the differential of $f$ at $p$.
Proof. Let $X_{p} \in T_{p} N$ be any tangent vector and $\ell \in C_{g f(p)}^{\infty}(P)$, then one has

$$
\begin{gathered}
\left((g f)_{*} X_{p}\right)(\ell)=X_{p}(\ell(g f)) \\
\left(\left(g_{*} f_{*}\right) X_{p}\right)(\ell)=\left(g_{*}\left(f_{*} X_{p}\right)\right)(\ell)=\left(f_{*} X_{p}\right)(\ell g)=X_{p}((\ell g) f)=X_{p}(\ell(g f))
\end{gathered}
$$

Where we used the fact that $f_{*} X_{p} \in T_{f(p)} M$. Therefore we conclude that $\left((g f)_{*} X_{p}\right)(\ell)=$ $\left(\left(g_{*} f_{*}\right) X_{p}\right)(\ell)$.

Corollary 23.4.5. Let $f: N \rightarrow M$ be a $C^{\infty}$-isomorphism and $p \in N$ be any point, then the $R$-linear map $f_{*}: T_{p} N \rightarrow T_{f(p)} M$ is an isomorphism of $R$-vector spaces.

Proof. Since $f$ is an isomorphism of manifolds, consider its inverse $f^{-1}: M \rightarrow N$. Using the chain rule, we find that

$$
\begin{gathered}
\operatorname{id}_{T_{p} N}=\left(\mathrm{id}_{N}\right)_{* p}=\left(f^{-1} f\right)_{* p}=f_{* f(p)}^{-1} f_{* p}, \\
\operatorname{id}_{T_{f(p)} M}=\left(\mathrm{id}_{M}\right)_{* f(p)}=\left(f f^{-1}\right)_{* f(p)}=f_{* f^{-1} f(p)} f_{* f(p)}^{-1}=f_{* p} f_{* f(p)}^{-1}
\end{gathered}
$$

This shows that $f_{* p}$ and $f_{* f(p)}^{-1}$ are inverses of each other-thus isomorphisms.
Corollary 23.4.6 (Dimension invariance). If $U \subseteq \mathbf{R}^{n}$ is an open set $C^{\infty}$-isomorphic to an open set $V \subseteq \mathbf{R}^{m}$, then $n=m$.

Proof. Let $f: U \rightarrow V$ be a $C^{\infty}$-isomorphism. For any $p \in U$ one has that $f_{* p}: T_{p} U \rightarrow$ $T_{f(p)} V$ is an isomorphism of vector spaces by Corollary 23.4.5. Since $\operatorname{dim}_{\mathbf{R}} T_{p} U=n$ and $\operatorname{dim}_{\mathbf{R}} T_{f(p)} V=m$, then necessarily $n=m$.
Definition 23.4.7 (Immersions and submersions). Let $f: N \rightarrow M$ be a $C^{\infty}$-morphism, and $p \in N$ be any point-also assume that $\operatorname{dim} N=n$ and $\operatorname{dim} M=m$. We define the following two concepts concerning the differential $f_{* p}: T_{p} N \rightarrow T_{f(p)} M$ :
(a) The map $f$ is said to be an immersion at $p$ if $f_{* p}$ is injective-if that is the case, then $n \leqslant m$. If $f$ is an immersion at every point, we simply classify $f$ as an immersion.
(b) The map $f$ is said to be an submersion at $p$ if $f_{* p}$ is surjective-if that is the case, then $m \leqslant n$. If $f$ is a submersion at every given point, we call $f$ a submersion.
Remark 23.4.8. Mind you: immersions do not need to be injective, neither do submersions need to be surjective.

We now restate Definition 23.2.4 and Definition 23.2.5 with this new terminology of immersions. Given a $C^{\infty}$-morphism $f: M \rightarrow N$, if $f$ is an immersion and is injective, then the pair $(M, f)$ defines a submanifold of $N$. Consequently, $f: M \rightarrow f M$ is a $C^{\infty}$ isomorphism, being an embedding. Later (see Definition 23.5.1) we shall define $f M$ to be an embedded submanifold on $N$.

## Tangent Space Basis at a Point

Lemma 23.4.9. Let $M$ be an $n$-dimensional smooth manifold and $(U, \phi)$ be a chart about a point $p \in M$, then for every $1 \leqslant j \leqslant n$ we have

$$
\phi_{*}\left(\left.\frac{\partial}{\partial \phi_{j}}\right|_{p}\right)=\left.\frac{\partial}{\partial \pi_{j}}\right|_{\phi(p)}
$$

where $\left(\pi_{j}\right)_{j=1}^{n}$ are the local coordinates of $\mathbf{R}^{n}$-that is, the canonical projections.
Proof. Let $f \in C_{\phi(p)}^{\infty}\left(\mathbf{R}^{n}\right)$ be any smooth morphism, then we have

$$
\left(\phi_{*}\left(\left.\frac{\partial}{\partial \phi_{j}}\right|_{p}\right)\right)(f)=\left.\frac{\partial}{\partial \phi_{j}}\right|_{p}(f \phi)=\left.\frac{\partial}{\partial \pi_{j}}\right|_{\phi(p)}(f \phi) \phi^{-1}=\left.\frac{\partial}{\partial \pi_{j}}\right|_{\phi(p)} f\left(\phi \phi^{-1}\right)=\left.\frac{\partial}{\partial \pi_{j}}\right|_{\phi(p)} f
$$

for each $1 \leqslant j \leqslant n$.
Proposition 23.4.10 (Tangent space basis at a point). Let $M$ be an $n$-dimensional smooth manifold, and $(U, \phi)$ be a chart about $p \in M$. Then the $R$-vector space $T_{p} M$ has a basis

$$
\left(\left.\frac{\partial}{\partial \phi_{1}}\right|_{p}, \ldots,\left.\frac{\partial}{\partial \phi_{n}}\right|_{p}\right)
$$

Proof. Via Lemma 23.4 .9 we know that the map $\phi_{*}: T_{p} M \rightarrow T_{\phi(p)} \mathbf{R}^{n}$ is $R$-linear and sends each $\left.\frac{\partial}{\partial \phi_{j}}\right|_{p}$ to the corresponding base element $\left.\frac{\partial}{\partial \pi_{j}}\right|_{\phi(p)}$ of $T_{\phi(p)} \mathbf{R}^{n}$. Therefore $\phi_{*}$ is an $\mathbf{R}$-linear isomorphism and the collection of tangent vectors $\left.\frac{\partial}{\partial \phi_{j}}\right|_{p}$ is a basis. $\quad$.

Proposition 23.4.11 (Transition map matrix). Let $(U, \phi)$ and $(V, \psi)$ be coordinate charts for an $n$-dimensional manifold $M$. Then one has on the intersection $U \cap V$, for each $1 \leqslant j \leqslant n$,

$$
\frac{\partial}{\partial \phi_{j}}=\sum_{i=1}^{n} \frac{\partial \psi_{i}}{\partial \phi_{j}} \frac{\partial}{\partial \psi_{i}}
$$

Proof. Let $p \in U \cap V$ be any point and define $\left[a_{i j}\right]_{i, j=1}^{n}$ be a change of basis matrix from $\left(\left.\frac{\partial}{\partial \psi_{j}}\right|_{p}\right)_{j=1}^{n}$ to $\left(\left.\frac{\partial}{\partial \phi_{j}}\right|_{p}\right)_{j=1}^{n}$-that is, the coefficients of the matrix must satisfy

$$
\frac{\partial}{\partial \phi_{j}}=\sum_{k=1}^{n} a_{k j} \frac{\partial}{\partial y_{k}}
$$

Since $\frac{\partial \psi_{i}}{\partial \psi_{j}}=\delta_{i j}$ then by applying the component $\psi_{i}$ to the above equation of tangent vectors, one has

$$
\frac{\partial \psi_{i}}{\partial \phi_{j}}=\sum_{k=1}^{n} a_{k j} \frac{\partial \psi_{i}}{\partial \psi_{k}}=\sum_{k=1}^{n} a_{k j} \delta_{i k}=a_{i j}
$$

Which proves the proposition.

Proposition 23.4.12 (Local expression for the differential of a map). Let $f: N \rightarrow M$ be a $C^{\infty}$-morphism, and consider a point $p \in N$-moreover, let $\operatorname{dim} N=n$ and $\operatorname{dim} M=m$. Given coordinate charts $(U, \phi)$ about $p$, and $(V, \psi)$ about $f(p)$, the differential $f_{* p}: T_{p} N \rightarrow T_{f(p)} M$ can be locally represented-that is, with respect to the bases $\left(\left.\frac{\partial}{\partial \phi_{j}}\right|_{p}\right)_{j=1}^{n}$ and $\left(\left.\frac{\partial}{\partial \psi_{i}}\right|_{f(p)}\right)_{i=1}^{m}$-by the matrix

$$
\left[\frac{\partial f_{i}}{\partial \phi_{j}}(p)\right]_{1 \leqslant j \leqslant n}
$$

where $f_{i}:=\psi_{i} f$ is the $i$-th component of $f$ with respect to $\psi$, for each $1 \leqslant i \leqslant m$.
Proof. Given the said coordinate charts, $f_{* p}$ will be completely determined by its image on the local basis of $T_{p} N$ given by $(U, \phi)$. That is, $f_{* p}$ is given by the matrix $\left[a_{i j}\right]_{i, j=1}^{n}$ whose coefficients satisfy

$$
f_{*}\left(\left.\frac{\partial}{\partial \phi_{j}}\right|_{p}\right)=\left.\sum_{k=1}^{m} a_{k j} \frac{\partial}{\partial \psi_{k}}\right|_{f(p)}
$$

Moreover, if we apply $\psi_{i}$ to the above equation concerning tangent vectors and obtain

$$
a_{i j}=\left(\left.\sum_{k=1}^{m} a_{k j} \frac{\partial}{\partial \psi_{k}}\right|_{f(p)}\right)\left(\psi_{i}\right)=f_{*}\left(\left.\frac{\partial}{\partial \phi_{j}}\right|_{p}\right)\left(\psi_{i}\right)=\left.\frac{\partial}{\partial \phi_{j}}\right|_{p}\left(\psi_{i} f\right)=\frac{\partial f_{i}}{\partial \phi_{j}}(p) .
$$

## Rank, \& Critical \& Regular points

Definition 23.4.13 (Rank at a point). Let $f: N \rightarrow M$ be a $C^{\infty}$-morphism, and $p \in N$ be any point. We define the rank of $f$ at the point $p$ to be

$$
\operatorname{rank}_{p} f:=\operatorname{rank} f_{* p}
$$

Locally, by the independence of the choice of coordinate charts-given a chart $(U, \phi)$ about $p$, and a chart $(V, \psi)$ about $f(p)$-one has

$$
\operatorname{rank}_{p} f=\operatorname{rank}\left[\frac{\partial f_{i}}{\partial \phi_{j}}(p)\right] .
$$

Definition 23.4.14 (Critical \& Regular). Let $f: N \rightarrow M$ be a $C^{\infty}$-morphism, and consider any two points $p \in N$ and $q \in M$. We define the following concerning the classification of the point $p$ :
(a) We say that $p$ is a critical point of $f$ if the differential $f_{* p}$ isn't surjective. The point $q$ is said to be a critival value if it is the image of a critical point.
(b) We say that $p$ is a regular point of $f$ if the differential $f_{*}$ is surjective-that is, $f$ is a submersion at $p$.

Finally, if $q$ is not a critical value, we say that it is a regular value-we do not impose $q$ to be an element of the image of $f$.

Proposition 23.4.15. Let $f: M \rightarrow \mathbf{R}$ be a $C^{\infty}$-morphism, and $p \in M$ be any point. Then $p$ is a critical point of $f$ if and only if there exists a chart $(U, \phi)$ about $p$ such that

$$
\frac{\partial f}{\partial \phi_{j}}(p)=0
$$

for all $1 \leqslant j \leqslant n$.
Proof. Let $(V, \psi)$ be any chart about $p$. Since $T_{f(p)} \mathbf{R} \simeq \mathbf{R}$, one has that the differential $f_{* p}: T_{p} M \rightarrow \mathbf{R}$ has rank $f_{* p}=1$-which means that $f_{* p}$ is surjective-or rank $f_{* p}=0$, that is, $f_{* p}=0$. Therefore $f_{* p}$ isn't surjective (i.e. $f$ isn't a submersion at $p$ ) if and only if $f_{* p}$ is zero-which is equivalent to the vanishing of all partial derivatives.

## Tangent Bundle

Given an $n$-manifold $M$, define

$$
T M:=\bigcup_{p \in M} T_{p} M
$$

which will be associated with a natural projection $\pi: T M \rightarrow M$ mapping each pair $(p, \xi) \in T M$, where $\xi \in T_{p} M$, to the point $p \in M$.

We wish to define a smooth structure on $T M$. To that end, let $p \in M$ be any point and $\phi: U \rightarrow U^{\prime} \subseteq \mathbf{R}^{n}$ be a chart of $M$ about $p$, forming a local basis $\left(\frac{\partial}{\partial \phi_{j}}\right)_{j=1}^{n}$ for the spaces $T_{x} M$ for any $x \in U$. From this we conclude that there exists a linear isomorphism $\pi^{-1} U \simeq U \times \mathbf{R}^{n} \simeq U^{\prime} \times \mathbf{R}^{n}$ associating each pair $\left(p, \xi=\sum_{j=1}^{n} a_{j} \frac{\partial}{\partial \phi_{j}}\right) \in T M$ to $\left(p, \phi_{*} \xi\right) \in U \times \mathbf{R}^{n}$ or $\left(\phi p, \phi_{*} \xi\right) \in U^{\prime} \times \mathbf{R}^{n}$. We shall take the latter, which has the form

$$
(\phi \pi) \times \phi_{*}: \pi^{-1} U \longrightarrow U^{\prime} \times \mathbf{R}^{n}
$$

as a chart on $T M$. To see that this gives a smoothly compatible structure to $T M$, let $\psi: V \rightarrow \mathbf{R}^{n}$ be another chart on $M$ and consider the transition map $\theta:=\psi \phi^{-1}: \phi(U \cap$ $V) \rightarrow \psi(U \cap V)$. The corresponding transition map between the charts $(\phi \pi) \times \phi_{*}$ and $(\psi \pi) \times \psi_{*}$ is then $\theta \times \theta_{*}$, which is smooth. This proves that the constructed atlas gives to $T M$ a smooth structure of a $2 n$-dimensional manifold. We shall refer to such manifold as the tangent bundle of $M$.

Definition 23.4.16 (Vector field). Let $M$ be a manifold. A vector field on $M$ is a section of the natural projection $\pi: T M \rightarrow M$-that is, a $C^{\infty}$-morphism $\xi: M \rightarrow T M$ for which $\pi \xi=\mathrm{id}_{M}$. More explicitly, given a point $p \in M$ and a chart $\phi$ of $M$ about $p$, we have

$$
\xi p=\sum_{j=1}^{n} \alpha_{j}(p) \frac{\partial}{\partial \phi_{j}}
$$

where $\alpha_{j}: M \rightarrow \mathbf{R}$ are $C^{\infty}$-morphisms. The collection of all vector fields on $M$ is denoted by $\mathfrak{X} M$.

Definition 23.4.17 (Flow). Let $M$ be an $m$-manifold. A smoothflow on $M$ is a smooth action of the abelian topological group $(\mathbf{R},+)$ on $M$, that is, a $C^{\infty}$-morphism $\theta: \mathbf{R} \times M \rightarrow M$ satisfying:

- For all $x \in M$, we have $\theta(0, x)=x$.
- For every pair $s, t \in \mathbf{R}$, and point $x \in M$, one has $\theta(s+t, x)=\theta(s, \theta(t, x))$.

Definition 23.4.18 (Tangent field to a flow). Let $\theta$ be a flow on a manifold $M$. We define a tangent field of $\theta$ to be the vector field $\xi: M \rightarrow T M$ given by

$$
\xi p:=\left.\theta_{*} \frac{\mathrm{~d}}{\mathrm{~d} t}\right|_{(0, p)}
$$

that is, the vector $\xi p \in T M$ is tangent to the curve $\gamma(t):=\theta(t, p)$ at the point $t=0$.
Example 23.4.19. Given a sphere $S^{2 n-1} \hookrightarrow \mathbf{C}^{n}$, consider the flow $\theta: \mathbf{R} \times S^{2 n-1} \rightarrow S^{2 n-1}$ given by

$$
\theta(t, z):=e^{i t} z
$$

The tangent field $\xi: S^{2 n-1} \rightarrow T S^{2 n-1}$ to $\theta$ is then

$$
\xi(z)=\left.\theta_{*} \frac{\mathrm{~d}}{\mathrm{~d} t}\right|_{t=0} z=\left.\frac{\mathrm{d}}{\mathrm{~d} t}\right|_{t=0} e^{i t} z=i z
$$

Therefore $\xi$ is a unitary vector field on the $(2 n-1)$-sphere.
Definition 23.4.20 (Parallelizable manifold). An $m$-manifold $M$ is said to be parallelizable if there exists a $C^{\infty}$-isomorphism $T M \xrightarrow{\simeq} M \times \mathbf{R}^{n}$ for which the restriction $T_{p} M \rightarrow p \times \mathbf{R}^{n}$ is an $\mathbf{R}$-linear isomorphism for each $p \in M$.

Example 23.4.21. The circle is parallelizable. Indeed, we can use $\xi: M \rightarrow T M$ as the tangent vector field to the flow $\mathbf{R} \times S^{1} \rightarrow S^{1}$ mapping $(t, z) \mapsto e^{i t} z$.

Proposition 23.4.22. The $n$-torus is parallelizable.
Proposition 23.4.23. If $n$ is odd, then $S^{n} \times S^{k}$ is parallelizable for all $k \geqslant 1$.
To prove: $n$-torus and product of spheres are parallelizable

### 23.5 Submanifolds

## Embedded Submanifolds

We shall give two descriptions of what is called an embedded submanifold:
Definition 23.5.1 (Embedded submanifold). A subspace $S \subseteq N$ is said to be an embedded submanifold on a manifold $N$ if the canonical inclusion $S \hookrightarrow N$ is a smooth embedding.

Definition 23.5.2 (Embedded submanifold, again ${ }^{2}$ ). Let $S \subseteq N$ be a subspace of the $n$-dimensional manifold $N$. We say that $S$ is a $k$-dimensional embedded submanifold of $N$ if for every $p \in S$ there exists a coordinate chart $(U, \phi)$ about $p$ such that $n-k$ coordinates of $\phi$ vanish at the intersection $U \cap S$. By the possible rearrangement of the indices, we may assume that

$$
\left.\phi\right|_{u \cap s}=\left(\phi_{1}, \ldots, \phi_{k}, 0, \ldots, 0\right) .
$$

The chart $(U, \phi)$ is then referred to as an adapted chart relative to $S$. We shall denote by $\phi_{S}: U \cap S \rightarrow \mathbf{R}^{k}$ the induced $C^{\infty}$-isomorphism given by

$$
\phi_{S}:=\left(\phi_{1}, \ldots, \phi_{k}\right),
$$

where $\left(U \cap S, \phi_{S}\right)$ is a chart for $S$ in the subspace topology.
The proof of the equivalence between Definition 23.5.1 and Definition 23.5.2 shall be postponed to Theorem 23.6.6.

Definition 23.5.3 (Codimension). Let $S \subseteq N$ be a $k$-dimensional submanifold of the $n$-dimensional manifold $N$. We define the codimension of $S$ as a submanifold of $N$ to be

$$
\operatorname{codim}_{N} S:=n-k
$$

Proposition 23.5.4 (Embedded submanifolds are manifolds). Let $S$ be a embedded submanifold of $N$ and let $\left(U_{\gamma}, \phi_{\gamma}\right)_{\gamma \in \Gamma}$ be a collection of adapted charts of $N$ relative to $S$, such that the family of charts cover $S$. Then the induced collection $\left(U_{\gamma} \cap S, \phi_{\gamma} S\right)_{\gamma \in \Gamma}$ is an atlas for $S$, and therefore $S$ is itself a manifold. If $\operatorname{dim} N=n$ and $S$ is locally defined by the vanishing of $n-k$ coordinates, then $\operatorname{dim} S=k$.

Proof. Consider two intersecting adapted charts $(U, \phi)$ and $(V, \psi)$ of $N$ contained in the above-mentioned collection. Let $p \in(U \cap V) \cap S$ be any point, then if $\phi(p)=$ $\left(x_{1}, \ldots, x_{k}, 0, \ldots, 0\right)$ and $\psi(p)=\left(y_{1}, \ldots, y_{k}, 0, \ldots, 0\right)$, we obtain

$$
\psi_{S} \phi_{S}^{-1}\left(x_{1}, \ldots, x_{k}\right)=\psi_{S}(p)=\left(y_{1}, \ldots, y_{k}\right)
$$

Since the natural projections $\pi_{j}\left(\psi_{S} \phi_{S}^{-1}\right):(U \cap V) \cap S \rightarrow \mathbf{R}$ are $C^{\infty}$-morphisms, then one concludes that the transition map $\psi_{S} \phi_{S}$ is itself a $C^{\infty}$-morphism, showing that the charts are compatible. Therefore the induced family of charts $\left(U_{\gamma} \cap S, \phi_{\gamma} s\right)_{\gamma \in \Gamma}$ is a smooth atlas for $S$. It is immediate that $S$ is a $k$-dimensional smooth manifold.

## Fibre of $C^{\infty}$-Morphisms

Definition 23.5.5 (Fibres \& levels). Given a $C^{\infty}$-morphism $f: N \rightarrow M$, we say that a point $q \in M$ is the level of the fibre $f^{-1} q$-in particular, if $q$ is a regular value, then we say that $f^{-1} q$ is a regular fibre.

[^33]Lemma 23.5.6. Let $g: N \rightarrow \mathbf{R}$ be a $C^{\infty}$-morphism, and $q \in \mathbf{R}$ be any regular value. If $f: N \rightarrow \mathbf{R}$ is defined by $f:=g-q$, then 0 is a regular value of $f$ and $g^{-1} q=f^{-1} 0$. That is, one can construct a smooth map such that its regular zero fibre equals to a given regular fibre of $g$.

Proof. From construction, it is obvious that $g^{-1} q=f^{-1} 0$. To show that 0 is a regular value of $f$, notice that for all $p \in N$ one has $f_{* p}=g_{* p}$-therefore $f$ and $g$ have the exact same critical points, and thus the same regular values. It follows that $f^{-1} 0$ is indeed a regular fibre of $f$.

Lemma 23.5.7. Let $g: N \rightarrow \mathbf{R}$ be a $C^{\infty}$-morphism, and let $q \in \operatorname{im} g$ be a regular value. Then the non-empty regular fibre $g^{-1} q$ is a regular submanifold of $N$ with codimension 1.

Proof. Assume that $\operatorname{dim} N=n$. As in Lemma 23.5.6, define $f: N \rightarrow \mathbf{R}$ to be the $C^{\infty}$ morphism given by $f:=g-q$, so that $g^{-1}(q)=f^{-1}(0)$ and 0 is a regular value of $f$. Given any $p \in f^{-1}(0)$, by definition we have that $p$ is a regular point of $f$ and by Proposition 23.4.15 we know that any chart $(U, \phi)$ about $p$ has a non-vanishing partial derivative $\frac{\partial f}{\partial \phi_{j}}(p) \neq 0$ for some index $j$. By permutation of the indices, one may assume without much disturbance that $\frac{\partial f}{\partial \phi_{1}}(p) \neq 0$ is a non-vanishing partial derivative. Consider the $C^{\infty}$-morphism $\Phi: U \rightarrow \mathbf{R}^{n}$ defined by $\Phi:=\left(f, \phi_{2}, \ldots, \phi_{n}\right)$, whose jacobian matrix equals

$$
\text { Jac } \Phi=\left[\begin{array}{ccccc}
\frac{\partial f}{\partial \phi_{1}} & \frac{\partial f}{\partial \phi_{2}} & \cdots & \frac{\partial f}{\partial \phi_{n-1}} & \frac{\partial f}{\partial \phi_{n}} \\
\frac{\partial \phi_{2}}{\partial \phi_{1}} & \frac{\partial \phi_{2}}{\partial \phi_{2}} & \cdots & \frac{\partial \phi_{2}}{\partial \phi_{n-1}} & \frac{\partial \phi_{2}}{\partial \phi_{n}} \\
\vdots & \vdots & \ddots & \vdots & \vdots \\
\frac{\partial \phi_{n}}{\partial \phi_{1}} & \frac{\partial \phi_{n}}{\partial \phi_{2}} & \cdots & \frac{\partial \phi_{n}}{\partial \phi_{n-1}} & \frac{\partial \phi_{n}}{\partial \phi_{n}}
\end{array}\right]=\left[\begin{array}{ccccc}
\frac{\partial f}{\partial \phi_{1}} & \frac{\partial f}{\partial \phi_{2}} & \cdots & \frac{\partial f}{\partial \phi_{n-1}} & \frac{\partial f}{\partial \phi_{n}} \\
0 & 1 & \ldots & 0 & 0 \\
\vdots & \vdots & \ldots & \ddots & \vdots \\
0 & 0 & 0 & \cdots & n
\end{array}\right]
$$

Which has determinant $\operatorname{det}(\operatorname{Jac} \Phi(p))=\frac{\partial f}{\partial \phi_{1}}(p) \neq 0$ at the point $p$. From the inverse map theorem (see Corollary 23.1.36) there must exist a neighbourhood $X \subseteq M$ of $p$ such that $(X, \Phi: X \rightarrow \Phi(X))$ is a chart about $p$ for $M$. Notice that by construction one has $\left.\Phi\right|_{X \cap f^{-1}(0)}=\left(0, \Phi_{2}, \ldots, \Phi_{n}\right)$, which shows that $(X, \Phi)$ is an adapted chart of $M$ relative to $f^{-1}(0)=g^{-1}(q)$. This shows that every point of $g^{-1}(q)$ has a corresponding adapted chart in $M$, ensuring that $g^{-1}(q)$ is a regular submanifold with $\operatorname{codim} g^{-1}(q)=1 \quad$ b

Theorem 23.5.8 (Regular fibres are embedded submanifolds). Let $f: N \rightarrow M$ be a $C^{\infty}$ morphism, and let $\operatorname{dim} N=n$ while $\operatorname{dim} M=m$. Given a regular value $q \in \operatorname{im} f$, the non-empty regular fibre $f^{-1} q$ is a embedded submanifold of $N$ with $\operatorname{dim} f^{-1} q=n-m$.

Proof. We could simply use Lemma 23.5.6, but we shall construct the proof from the ground level. Let $(V, \psi)$ be a chart of $M$ about $q$ satisfying $\psi q=0 \in \mathbf{R}^{m}$ (that is, $(V, \psi)$ is centred at $q$ )-so that $f^{-1} V$ is a neighbourhood of $f^{-1} q$ in $N$, and one has

$$
f^{-1} V \cap f^{-1} q=f^{-1} q=\left(\left.\psi f\right|_{f^{-1} V}\right)^{-1} 0
$$

Therefore $f^{-1} q$ is the zero fibre of the map $\psi f$, which we can analyse locally as the zero fibres of the collection of local projections $\left(f_{j}=\pi_{j}\left(\left.\psi f\right|_{f^{-1} V}\right)\right)_{j=1}^{m}$. Since $q \in \operatorname{im} f$ is a regular value, every point $p \in f^{-1} q$ is regular and thus has a surjective $f_{* p}$-that is, $n \geqslant m$ and rank $f_{* p}=m$. Let $(U, \phi)$ be a chart about $p$ in $N$ where $U \subseteq f^{-1} V$. Consider the local Jacobian representation of $f_{* p}$ at $(U, \phi)$, and assume that the first $m \times m$ block of $\left[\frac{\partial f_{i}}{\partial \phi_{j}} p\right]$ is invertible-if that is not the case, simply rearrange the indices so that this is true.

Define a $C^{\infty}$-morphism $\phi^{\prime}: U \rightarrow \mathbf{R}^{n}$ by making $\phi_{j}^{\prime}:=f_{j}=\pi_{j}\left(\left.\psi f\right|_{f^{-1} V}\right)$ for $1 \leqslant j \leqslant m$, while $\phi_{j}^{\prime}:=\phi_{j}$ for the remaining $m<j \leqslant n$. To show that $\phi^{\prime}$ is locally invertible at $p$, notice that the local Jacobian matrix of $f$ about $\phi^{\prime}$ at $p$ is

$$
\left.\mathrm{Jac}_{\phi^{\prime}} f\right|_{p}=\left[\begin{array}{ll}
\frac{\partial f_{i}}{\partial \phi_{j}} p & \frac{\partial f_{i}}{\partial \phi_{\beta}} p \\
\frac{\partial \phi_{\alpha}}{\partial \phi_{j}} p & \frac{\partial \phi_{\alpha}}{\partial \phi_{\beta}} p
\end{array}\right]=\left[\begin{array}{cc}
\frac{\partial f_{i}}{\partial \phi_{j}} p & \frac{\partial f_{i}}{\partial \phi_{\beta}} p \\
0 & \text { id }
\end{array}\right]
$$

where $1 \leqslant i, j \leqslant m$ and $m<\alpha, \beta \leqslant n$. This shows that

$$
\operatorname{det}\left(\left.\operatorname{Jac}_{\phi^{\prime}} f\right|_{p}\right)=\operatorname{det}\left[\frac{\partial f_{i}}{\partial \phi_{j}} p\right]_{1 \leqslant i, j \leqslant m} \neq 0
$$

that is, $\phi^{\prime}$ is locally invertible at $p$ and by Corollary 23.1.36 there exists $U_{p} \subseteq U$ such that $\phi^{\prime}: U_{p} \rightarrow \phi^{\prime}\left(U_{p}\right)$ is a $C^{\infty}$-isomorphism, and thus $\left(U, \phi^{\prime}\right)$ is a chart about $p$ in $N$. Notice that, by construction, $f^{-1} q$ is the set obtained by imposing $\phi_{j}^{\prime}=f_{j}=0$ for each $1 \leqslant j \leqslant m$-thus $\left(U_{p}, \phi^{\prime}\right)$ is an adapted chart for $N$ relative to $f^{-1} q$. Therefore $f^{-1} q$ admits an adapted chart for each of its points, showing that it is a embedded submanifold of dimension $n-m$.

Corollary 23.5.9 (Producing adapted charts). Let $f: N \rightarrow \mathbf{R}^{m}$ be a $C^{\infty}$-morphism and $\operatorname{dim} N=n$, consider the zero fibre $f^{-1} 0$. If there exists a chart $(U, \phi)$ about $p \in f^{-1} 0$ such that

$$
\operatorname{det}\left[\frac{\partial f_{i}}{\partial \phi_{j}} p\right]_{1 \leqslant i, j \leqslant m} \neq 0
$$

then there exists a neighbourhood $U_{p} \subseteq N$ of $p$ such that the pair $\left(U_{p}, \phi^{\prime}\right)$, where

$$
\phi^{\prime}:=\left(f_{1}, \ldots, f_{m}, \phi_{m+1}, \ldots, \phi_{n}\right): U_{p} \longrightarrow \phi^{\prime} U_{p}
$$

is a $C^{\infty}$-isomorphism, forms an adapted chart for $N$ relative to $f^{-1} 0$.
Lemma 23.5.10. Let $M$ be a manifold and $S \subseteq M$ be a subset. Then $S$ is a embedded submanifold of $M$ with codimension $k$ if and only if every point of $S$ has a neighbourhood $U \subseteq M$ such that $U \cap S$ is the regular zero fibre of a $C^{\infty}$-isomorphism $\phi: U \rightarrow \mathbf{R}^{k}$.

Proof. Let $\operatorname{dim} M:=m .(\Rightarrow)$ Assume $S$ is a regular submanifold. For any $p \in S$, take $(U, \phi)$ to be an adapted chart of $M$ with respect to $S$, so that

$$
\left.\phi\right|_{S}=\left(\phi_{1}, \ldots, \phi_{m-k}, 0, \ldots, 0\right) .
$$

Define $\widehat{\phi}: U \rightarrow \mathbf{R}^{k}$ to be the $C^{\infty}$-isomorphism given by $\widehat{\phi}:=\left(\phi_{m-k+1}, \ldots, \phi_{m}\right)$, so that one has

$$
S \cap U=\widehat{\phi}^{-1} 0
$$

$(\Leftarrow)$ For the converse, suppose that $S$ admits a pair $\left(U, \widehat{\phi}: U \rightarrow \mathbf{R}^{k}\right)$ about each point $p \in S$ such that $U \cap S$ is the regular zero fibre of the $C^{\infty}$-isomorphism $\widehat{\phi}$. Therefore, any $C^{\infty}$-isomorphism $\phi: U \rightarrow \mathbf{R}^{m}$ such that

commutes, forms an adapted chart $(U, \phi)$ relative to $S$ about $p$. This shows that $S$ is a relative submanifold of $M$.

### 23.6 Rank of $C^{\infty}$-Morphisms

Theorem 23.6.1 (Constant rank, then the fibre is a embedded submanifold). Let $f: N \rightarrow$ $M$ be a $C^{\infty}$-morphism and $q \in M$ be any point. If $f$ has has a locally constant rank $k$ at $f^{-1} q$, then the fibre $f^{-1} q$ is a embedded submanifold of $N$ of codimension $k$.

Proof. Take any $p \in f^{-1} q$ and, by the constant rank theorem, let $(U, \phi)$ be a chart centred at $p$ and $(V, \psi)$ be a chart centred at $f p=q$ such that

$$
\psi f \phi^{-1}\left(x_{1}, \ldots, x_{n}\right)=\left(x_{1}, \ldots, x_{k}, 0, \ldots, 0\right) \in \mathbf{R}^{m}
$$

The zero fibre $\left(\psi f \phi^{-1}\right)^{-1} 0$ is therefore the common zero fibre of the coordinates $\bigcap_{j=1}^{k}\left(\psi f \phi^{-1}\right)_{j}^{-1} 0$. Notice that

$$
\phi\left(f^{-1} q\right)=\phi\left(f^{-1}\left(\psi^{-1} 0\right)\right)=\left(\psi f \phi^{-1}\right)^{-1} 0
$$

therefore $U \cap f^{-1} q=\bigcap_{j=1}^{k} \phi_{j}^{-1} 0$, proving that $(U, \phi)$ is an adapted chart of $N$ about $p$ with respect to $f^{-1} q$.

Theorem 23.6.2 (Constant rank). Let $N$ and $M$ be manifolds of dimensions $n$ and $m$, respectively. Given a $C^{\infty}$-morphism $f: N \rightarrow M$ with local constant rank $k$ in a neighbourhood of a $p \in N$, there are charts $(U, \phi)$ centred at $p$ and $(V, \phi)$ centred at $f p$ such that

$$
\left.\pi_{j}\left(\psi f \phi^{-1}\right)\right|_{\phi U}=0
$$

for all $k<j \leqslant n$.
Proof. Let $\left(U^{\prime}, \phi^{\prime}\right)$ be a chart about $p$, and $\left(V^{\prime}, \psi^{\prime}\right)$ be a chart about $f p$, and consider the $C^{\infty}$-morphism

$$
\psi^{\prime} f \phi^{\prime-1}: \phi^{\prime}\left(f^{-1} V \cap U\right) \longrightarrow \mathbf{R}^{m}
$$

Since both $\phi^{\prime}$ and $\psi^{\prime}$ are $C^{\infty}$-isomorphism, it follows that $\psi^{\prime} f \phi^{\prime-1}$ has a local constant rank $k$ at $\phi^{\prime} p$. By Theorem A.5.11 we conclude that there exists a $C^{\infty}$-isomorphism $\phi^{\prime \prime}$ of a neighbourhood of $\phi^{\prime} p \in \mathbf{R}^{n}$ and a $C^{\infty}$-isomorphism $\psi^{\prime \prime}$ of a neighbourhood of $\psi f p \in \mathbf{R}^{m}$ for which

$$
\psi^{\prime \prime}\left(\psi^{\prime} f \phi^{\prime-1}\right) \phi^{\prime \prime-1}\left(x_{1}, \ldots, x_{n}\right)=\left(x_{1}, \ldots, x_{k}, 0, \ldots, 0\right)
$$

Therefore by letting $\phi:=\phi^{\prime \prime} \phi^{\prime}$ and $\psi:=\psi^{\prime \prime} \psi^{\prime}$ we obtain the desired result.
The following theorems are mere corollaries of the constant rank theorem: given a smooth map $f: M \rightarrow N$, if $f$ is an immersion at a point $p$, then it has a locally constant rank $m$ near $p$, on the other hand, if $f$ is a submersion at a point $q$, it has a locally constant rank $n$ near $q$.

Theorem 23.6.3 (Immersion). Let $\theta: M \rightarrow N$ be a $C^{\infty}$-morphism, and $\operatorname{dim} M=m$ while $\operatorname{dim} N=n$. If $\theta$ is an immersion at $p \in M$, then there are charts $\phi$ about $p$ and $\psi$ about $\theta p$ such that the diagram

commutes in Man. That is, $\theta$ locally acts as an inclusion about $p$.
Theorem 23.6.4. Let $\theta: M \rightarrow N$ be a $C^{\infty}$-morphism. If $\theta$ is a submersion at some point $p \in M$, there are charts $\phi$ about $p$ and $\psi$ about $\theta p$ for which

commutes in Man. That is, $\theta$ locally acts as a projection about $p$.
Corollary 23.6.5. A smooth submersion is an open map.
Proof. Let $f: M \rightarrow N$ be a smooth submersion and $U \subseteq M$ be an open subset. Let $f p \in f U$ be any point for which $p \in U$. Applying the submersion theorem to $f$, we can find a neighbourhood $V \subseteq U$ of $p$ for which $\left.f\right|_{V}$ acts as a projection, which implies that $f V \subseteq f U$ is a neighbourhood for $p$-since projections are open morphisms. This proves that $f U$ is an open set.

Theorem 23.6.6. Let $N$ be an $n$-manifold, and $M$ be an $m$-manifold. Then:
(a) If $f: N \rightarrow M$ is a smooth embedding, then the image $f N$ is an embedded submanifold of $M$ in the sense of Definition 23.5.2.
(b) If $N$ is an embedded submanifold of $M$ in the sense of Definition 23.5.2, then the canonical inclusion $t: N \hookrightarrow M$ is a smooth embedding-thus $N$ is an embedded submanifold in the sense of Definition 23.5.1.

This shows that the two definitions for embedded submanifolds are equivalent.
Proof. (a) By the immersion theorem, let $(U, \phi)$ be a chart centred at $p$ and $(V, \psi)$ be a chart centred at $f p$ such that

commutes-where $\phi U \subseteq \mathbf{R}^{n}$ and $\phi V \subseteq \mathbf{R}^{m}$. Then $f U$ is simply the common zero fibres of the last $m-n$ coordinates of $\psi$, that is, $f U=\bigcap_{j=n+1}^{m} \psi_{j}^{-1} 0$. Since $f$ is injective, it follows that $f N$-endowed with the subspace topology-is isomorphic to $N$. Therefore $f U$ is open in $f N$, thus there must exist an open set $V^{\prime} \subseteq M$ such that $V^{\prime} \cap f N=f U$ and hence $V^{\prime}$ is a neighbourhood of $f p$. Since

$$
\left(V \cap V^{\prime}\right) \cap f N=V \cap f U=f U
$$

it follows that $\left(V \cap V^{\prime}, \psi\right)$ is an adapted chart of $M$ about $f p$ with respect to $f N$. Therefore $f N$ is an embedded submanifold in the sense of Definition 23.5.2.
(b) Recall that both $\iota N$ and $N$ are endowed with the subspace topology. Since $\iota$ is injective, the induced map $\iota: N \rightarrow \iota N$ is a topological isomorphism—therefore $\iota$ is a topological embedding. In order to show that $\iota$ is an immersion, let $p \in N$ be any point, and take an adapted chart $(V, \psi)$ of $M$ about $p$ with respect to $N$-that is, $V \cap N=\bigcap_{j=n+1}^{m} \psi_{j}^{-1} 0$. Define $\psi_{N}:=\left(\psi_{1}, \ldots, \psi_{n}\right): V \cap N \rightarrow \psi_{N}(V \cap N) \subseteq \mathbf{R}^{n}$, then

commutes, showing that $\iota$ is an immersion.

### 23.7 Transversality

Definition 23.7.1 (Transversal $C^{\infty}$-morphism). Let $M$ be a manifold and $S \subseteq M$ be an embedded submanifold. A $C^{\infty}$-morphism $f: N \rightarrow M$ is said to be transversal to $S$, denoted $f \pitchfork S$, if for every $p \in f^{-1} S$ one has

$$
f_{*}\left(T_{p} N\right)+T_{f p} S=T_{f p} M
$$

Moreover, we can equally define this notion for embedded submanifolds: let $N$ and $L$ be embedded submanifolds of $M$. We say that $N$ intersects $L$ transversely, denoted $N \pitchfork L$, if for any $p \in N \cap L$ we have

$$
T_{p} N+T_{p} L=T_{p} M
$$

Theorem 23.7.2 (Transversality). Let $M$ be a manifold, and $S \subseteq M$ be a embedded submanifold with $\operatorname{codim}_{M} S=k$. Then every $C^{\infty}$-morphism $f: N \rightarrow M$ that is transversal to $S$ is such that $f^{-1} S$ is a embedded submanifold of codimension $k$ in $N$.

Proof. Let $\operatorname{dim} M:=m$ and $\operatorname{dim} N:=n$. Since $S$ is a regular submanifold of $M$, given any point $p \in f^{-1} S$, we can find an adapted chart $(U, \phi)$ of $M$ with respect to $S$ that is centred at $f p \in S$ and such that

$$
U \cap S=\left\{q \in M: \phi_{j} q=0 \text { for each } m-k<j \leqslant m\right\}
$$

that is, $U \cap S$ is the zeros of the last $m-k$ coordinates of $\phi$.
Define $g: U \rightarrow \mathbf{R}^{k}$ as the $C^{\infty}$-isomorphism given by those last $m-k$ coordinates:

$$
g:=\left(\phi_{m-k+1}, \ldots, \phi_{m}\right),
$$

so that $\left.g\right|_{S}=0$. Since $f p \in U$ then $f^{-1} U \cap f^{-1} S$ is a neighbourhood for $p$ in $N$, and $p \in\left(\left.g f\right|_{f^{-1} U}\right)^{-1} 0$ because $f p \in S$, therefore

$$
f^{-1} U \cap f^{-1} S=\left(\left.g f\right|_{f^{-1} U}\right)^{-1} 0
$$

We shall prove that 0 is a regular value of $\left.g f\right|_{f^{-1} U}$ so that $f^{-1} U \cap f^{-1} S$ is a regular zero fibre of the map. To that end, let $z \in T_{0} \mathbf{R}^{k}$ and $x \in\left(\left.g f\right|_{f^{-1} U}\right)^{-1} 0$ be any two elements. Since $g$ is an isomorphism, 0 is a regular value of $g$, therefore there exists $y \in T_{f x} M$ for which $g_{* f x}(y)=z$. Using the fact that $f$ is transversal to $S$, there exists $y_{0} \in T_{f x} S$ and $v \in T_{x} N$ for which $y=y_{0}+f_{* x} v$. Recall that $\left.g\right|_{S}=0$ from construction, therefore $g_{*} f x\left(y_{0}\right)=0$. Hence one has

$$
z=g_{* f x} y=g_{* f x}\left(y_{0}+f_{* x} v\right)=g_{* f x} f_{* x} v=(g f)_{* x} v,
$$

showing that $(g f)_{* x}$ is surjective-hence $x$ is a regular point of $\left.g f\right|_{f^{-1} U}$. This shows that 0 is a regular value of $\left.g f\right|_{f^{-1} U}$, therefore the subset $f^{-1} U$ is a neihbourhood of $p$ satisfying the desired condition specified in Lemma 23.5.10-moreover, since $p$ was any point of $f^{-1} S$, the theorem follows by just mentioned lemma.

Theorem 23.7.3. Let $M$ be an $m$-manifold, and both $N$ and $L$ be submanifolds-whose dimensions are $n$ and $\ell$, respectively. If $N \pitchfork L$ in $M$, then $N \cap L$ is a submanifold of $M$ with dimension $\operatorname{dim} N \cap L=n+\ell-m$.

Proof. Let $p \in N \cap L$ be any point. Let $(U, \phi)$ and $(V, \psi)$ be adapted charts of $M$ about $p$ with respect to $N$ and $L$. Let $W:=U \cap V$ and define the induced maps

$$
\eta:=\left(\phi_{n+1}, \ldots, \phi_{m}\right): W \longrightarrow \mathbf{R}^{m-n} \quad \text { and } \quad \lambda:=\left(\psi_{\ell+1}, \ldots, \psi_{m}\right): W \longrightarrow \mathbf{R}^{m-\ell}
$$

so that $W \cap N=\eta^{-1} 0$ and $W \cap L=\lambda^{-1} 0$ are regular fibres of $\eta$ and $\lambda$, respectively.
Consider the product map $\eta \times \lambda: W \rightarrow \mathbf{R}^{m-n} \times \mathbf{R}^{m-\ell}$ —we wish to show that ( $\eta \times$ $\lambda)^{-1} 0=W \cap(N \cap L)$ is a regular fibre of $\eta \times \lambda$, so that $(W, \phi \times \psi)$ is an adapted chart of $M$ about $p$ with respect to $N \cap L$. To that end, we shall consider the differential $(\eta \times \lambda)_{*}: T_{p} M \rightarrow \mathbf{R}^{m-n} \times \mathbf{R}^{m-\ell}$, whose kernel is

$$
\operatorname{ker}(\eta \times \lambda)_{*}=\operatorname{ker} \eta_{*} \cap \operatorname{ker} \lambda_{*}=T_{p} N \cap T_{p} L
$$

On the other hand, we know that

$$
\operatorname{dim} \operatorname{ker}(\eta \times \lambda)_{*}=\operatorname{dim} T_{p} N+\operatorname{dim} T_{p} L-\operatorname{dim} T_{p} M=n+\ell-m
$$

therefore by the rank plus nullity theorem one has

$$
\operatorname{rank}(\eta \times \lambda)_{*}=m-(n+\ell-m)=2 m-n-\ell=\operatorname{dim}\left(\mathbf{R}^{m-n} \times \mathbf{R}^{m-\ell}\right)
$$

which shows that $(\eta \times \lambda)_{*}$ is an epimorphism at $p$-thus 0 is a regular value of $\eta \times \lambda$. This concludes the proof that $N \cap L$ is an embedded submanifold of $M$ with dimension $n+\ell-m$.

### 23.8 Classification of 1-Manifolds

Definition 23.8.1 (Parametrization by arc-length). Let $I \subseteq \mathbf{R}$ be an interval, and $M$ be a smooth manifold. We say that a $C^{\infty}$-morphism $f: I \rightarrow M$ is a parametrization by arc-length if the restriction $f: I \rightarrow f I$ is a $C^{\infty}$-isomorphism, and if the velocity vector $f_{* s} 1 \in T_{f_{s}} M$ is unitary for each $s \in I$.

Lemma 23.8.2. Consider a pair of arc-length parametrizations $I \stackrel{f}{\rightarrow} M \stackrel{g}{\leftarrow} J$. Then $f I \cap g J$ has at most two connected components. One has the following properties concerning the number of connected components:
(a) If $f I \cap g J$ has only one connected component, then one can extend $f$ to an arc-length parametrization of $f I \cup g J$.
(b) If $f I \cap g J$ has two components, then $M \simeq S^{1}$.

Proof. Consider the $C^{\infty}$-isomorphism

$$
g^{-1} f: f^{-1}(f I \cap g J) \xlongequal{\leftrightharpoons} g^{-1}(f I \cap g J),
$$

sending open sets of $I$ to open sets of $J$, and with derivative $\pm 1$ everywhere-from the definition of the parametrization. Let $\Gamma$ be the pullback of the pair $(f, g)$-that is, composed of pairs $(s, t) \in I \times J$ such that $f s=g t$. Therefore $\Gamma$ is a closed subset of $I \times J$, with the product topology, and consists of line segments with slope $\pm 1$ by the behaviour of $g^{-1} f$. Since $g^{-1} f$ is an isomorphism, it must be the case that the line segments don't end abruptly, but extend from edge to edge of $I \times J$. From the fact that
$g^{-1} f$ is injective and constant derivative, it must be the case that $\Gamma$ is composed of at most two line segments-furthermore, if $\Gamma$ has two components, then they must have the same slope and the start and end edges has to be distinct. We analyse the number of components of $\Gamma$ :

- If $\Gamma$ has a single component, then one can extend $g^{-1} f$ to a line $\ell: \mathbf{R} \rightarrow \mathbf{R}$. Therefore the map $g \ell: I \cup \ell^{-1} J \rightarrow f I \cup g J$ forms an extension of $f$.
- If $\Gamma$ has two components, let those be the line segments connecting $(a, \alpha) \rightarrow(b, \beta)$ and $(c, \gamma) \rightarrow(d, \delta)$-whose points lie in $I \times J$, with ends in the boundary of the square, that is, $I=(a, d)$ and $J=(\gamma, \delta)$ for instance. By merely a translation of $J$, we may assume that $\gamma=c$ and $\delta=d$, for which we obtain the relations

$$
a<b \leqslant c<d \leqslant \alpha<\beta
$$

We may define a continuous maps $\theta:[a, \beta] \rightarrow \mathbf{R}$ given by $t \mapsto \frac{2 \pi t}{\alpha-a}$, and $h: S^{1} \rightarrow M$ mapping

$$
h(\cos (\theta t), \sin (\theta t)):= \begin{cases}f t, & \text { if } a<t<d \\ g t, & \text { if } c<t<\beta\end{cases}
$$

which is well defined since $f$ and $g$ agree on $[c, d]$ due to the translation of $J$. Since $h S^{1}$ is a compact open set of $M$, it follows that it must be the case that $h S^{1}=M$. Therefore, since the restrictions of $f$ and $g$ are $C^{\infty}$-isomorphisms, it follows that $h$ is a $C^{\infty}$-isomorphism.

Theorem 23.8.3 (Classification of 1-manifolds). Any smooth connected 1-manifold M is $C^{\infty}$-isomorphic to either the circle $S^{1}$ or to some interval of real numbers: $[0,1],[0,1)$, $(0,1]$ or $(0,1)$. In fact, the following is a complete classification list:
(1) If $M$ is compact without boundary, then $M$ is $C^{\infty}$-isomorphic to a circle.
(2) If $M$ is compact with boundary, then $M$ is $C^{\infty}$-isomorphic to a closed interval.
(3) If $M$ is non-compact without boundary, then $M$ is $C^{\infty}$-isomorphic to an open interval.
(4) If $M$ is non-compact with boundary, then $M$ is $C^{\infty}$-isomorphic to a half-open interval.

Proof. Given a parametrization by arc-length $f^{\prime}: J \rightarrow M$, via Lemma 23.8.2 one can extend $f^{\prime}$ to a maximal arc-length parametrization $f: I \rightarrow M$-so that $I$ is the maximal interval $f$ can be extended to.

Assuming $M \neq S^{1}$, suppose that there exists a limit point $x$ of $f I$ with $x \in M \backslash f I$, so that $f$ is not surjective. Let $U$ be a neighbourhood of $x$ and $g: I^{\prime} \rightarrow U$ be an arc-length parametrization of $U$. From Lemma 23.8 . 2 we can use $g$ to extend $f$ to a parametrization $f I \cup g J$, contradicting the hypothesis of maximality of $f$.

Corollary 23.8.4. The boundary of a compact 1-manifold has an even number of points.

Proof. Indeed, if $M$ is a compact 1-manifold with boundary, then it's isomorphic to the disjoint union of a collection of closed intervals, each of which has two boundary points, thus $M$ has an even number of boundary points.

## Chapter 24

## Cobordism

### 24.1 Cobordisms

## Unoriented Cobordisms

Definition 24.1.1 (Unoriented cobordism). Given a pair $\Sigma_{0}$ and $\Sigma_{1}$ of smooth compact ( $n-1$ )-manifolds without boundary, we define a cobordism between $\Sigma_{0}$ and $\Sigma_{1}$ to be a smooth compact $n$-manifold $M$ whose boundary is $\partial M=\Sigma_{0} \amalg \Sigma_{1}$. We thus call the manifolds $\Sigma_{0}$ and $\Sigma_{1}$ cobordant.

Example 24.1.2. Two interesting cobordisms are formed from the empty manifold to the circle, which shall be called birth-of-a-circle, and from the circle to the empty manifold, so called death-of-a-circle.

Lemma 24.1.3 (Cobordant zero and one dimensional manifolds). Two given compact 0 -manifolds without boundary are cobordant if and only if they have the same number of points modulo 2. Moreover, any two compact 1-manifolds without boundary are cobordant.

Proof. Let's consider the case of a pair of 0-manifolds $\Sigma_{0}$ and $\Sigma_{1}$. Notice that since every pair of points can be connected by a smooth curve, and every 1-manifold with boundary has an even number of boundary points ${ }^{1}$, it follows that $\Sigma_{0}$ and $\Sigma_{1}$ are cobordant if and only if the disjoint union $\Sigma_{0} \amalg \Sigma_{1}$ has an even number of points.

For the second statement, one should recall that a compact 1-manifold is the disjoint union of circles. Then we can choose one of the manifolds to attach copies of the death-of-a-circle cobordism for each of its circles, and attach birth-of-a-circle cobordisms for each of its respective circles of the other 1-manifold. This construction yields a cobordism between them.

[^34]
## Oriented Cobordisms

Consider the following setup: let $\Sigma$ be a closed submanifold of $M$ with codimension 1 , where $\operatorname{dim} M=n$. Assume both manifolds to be oriented.

Definition 24.1.4 (Positive normal). Let $\left[v_{1}, \ldots, v_{n-1}\right]$ be a positive basis for $T_{x} \Sigma$ for any given point $x \in \Sigma$. We say that a tangent vector $v \in T_{x} M$ is a positive normal if the induced basis $\left[v_{1}, \ldots, v_{n-1}, v\right]$ for $T_{x} M$ is positive.

Definition 24.1.5 (In and out boundaries). If $\Sigma$ is a connected component of $\partial M$, we call $\Sigma$ an in-boundary if a positive normal points inwards relative to $M$, and otherwise an out-boundary-when a positive normal points outward relative to $M$.

The notion of an in and out boundary allows us to define the notion of an oriented cobordism. From now on, a cobordism will always mean an oriented one, unless stated otherwise.

Definition 24.1.6 (Oriented cobordism). Let $\Sigma_{\text {in }}$ and $\Sigma_{\text {out }}$ be compact $(n-1)$-manifolds without boundary. We define an oriented cobordism between them to be a triple ( $M, \iota_{\text {in }}, l_{\text {out }}$ ), where $M$ is a smooth compact oriented $n$-manifold, and arrows

$$
\Sigma_{\text {in }} \xrightarrow{l_{\text {in }}} M \stackrel{l_{\text {out }}}{\longleftrightarrow} \Sigma_{\text {out }}
$$

which are $C^{\infty}$-isomorphisms when restricted to the in and out boundary of $M$, respectively. We shall denote the oriented cobordism $M$ as an arrow $M: \Sigma_{\text {in }} \Rightarrow \Sigma_{\text {out }}$.

Definition 24.1.7 (Equivalence of cobordisms). Given two cobordisms

$$
\Sigma_{\text {in }} \underset{N}{\stackrel{M}{\Longrightarrow}} \Sigma_{\text {out }}
$$

we say that $M$ is equivalent to the cobordism $N$ if there exists an orientation-preserving $C^{\infty}$-isomorphism $\phi: M \xrightarrow{\simeq} N$ such that the following diagram commutes in Man:


### 24.2 Elements of Morse Theory

Definition 24.2.1. Let $f: M \rightarrow I$ be a $C^{\infty}$-morphism, and $p \in M$ be a critical point of $f$. We call $p$ a non-degenerate point if there exists a chart about $p$ for which the local Hessian of $f$ is invertible. Furthermore, define the index of $f$ at $p$ to be the number of negative eigenvalues of the local Hessian.

Definition 24.2.2 (Morse maps). Given a smooth manifold $M$, we say that a $C^{\infty}$ _ morphism $f: M \rightarrow I$ is a Morse map if every critical point of $f$ is non-degenerate. If it happens to be the case that $M$ is a manifold with boundary, we shall require that $f^{-1} \partial I=\partial M$ and that the boundary points $\partial I=\{0,1\}$ are regular values of $f$-this ensures that $\partial M$ contains no critical points.

The existence of Morse maps is ensured by the following theorem:
Theorem 24.2.3. For any manifold $M$ and integer $2 \leqslant r \leqslant \infty$, the collection of Morse maps $M \rightarrow I$ is dense in $C^{r}(M, I)$.

The following is a generalization of the construction of attaching spaces:
Definition 24.2.4 (Gluing). Let $f: X \rightarrow Y$ and $g: X \rightarrow Z$ be topological morphisms. We define the gluing of $Y$ and $Z$ along $X$ to be the pushout


Explicitly, $Y \amalg_{X} Z$ is the quotient space of $Y \amalg Z$ where $y \sim z$ if and only if there exists a common $x \in X$ such that $f x=y$ and $g x=z$.

## Part VIII

## Analysis

## Chapter 25

## Holomorphic Functions

Definition 25.0.1 (Holomorphic function). A function is said to be holomorphic (also called analytic) if it is a differentiable complex-valued function with a complex variable.

Let $f: \mathbf{C} \rightarrow \mathbf{C}$ be a holomorphic function such that $f=u+\mathrm{i} v$ where $u, v: \mathbf{C} \rightrightarrows \mathbf{R}$. Let's write the variables of $f$ as $z=x+\mathrm{i} y$ so that $f$ may be seen as a function of two variables $\mathbf{R}^{2} \rightarrow \mathbf{R}^{2}$ given by $(x, y) \mapsto(u(x), v(y))$ from the bijection $\mathbf{C} \simeq \mathbf{R}^{2}$. Since $f$ is differentiable, then so are $u$ and $v$. Let's consider the definition of the derivative of $f$ at a point $z \in \mathbf{C}$ :

$$
\begin{equation*}
f^{\prime}(z)=\lim _{\mathrm{R} \ni t \rightarrow 0} \frac{f(z+t)-f(z)}{t}=\lim _{\mathrm{R} \ni t \rightarrow 0} \frac{f(z+\mathrm{i} t)-f(z)}{\mathrm{i} t} . \tag{25.1}
\end{equation*}
$$

If we consider the real case where we approach from the purely real values we get

$$
\begin{aligned}
\lim _{\mathbf{R} \exists t \rightarrow 0} \frac{f(z+t)-f(z)}{t} & =\lim _{t \rightarrow 0} \frac{u(x+t, y)+\mathrm{i} v(x+t, y)-(u(x, y)+\mathrm{i} v(x, y))}{t} \\
& =\lim _{t \rightarrow 0} \frac{u(x+t, y)-u(x, y)}{t}+\mathrm{i} \lim _{t \rightarrow 0} \frac{v(x+t, y)-v(x, y)}{t} \\
& =\frac{\partial u}{\partial x}+\mathrm{i} \frac{\partial v}{\partial x} .
\end{aligned}
$$

On the other hand, if we restrict the limit by approaching from strictly imaginary values, we obtain:

$$
\begin{aligned}
\lim _{\mathrm{R} \ni \rightarrow 0} \frac{f(z+\mathrm{i} t)-f(z)}{\mathrm{i} t} & =\lim _{t \rightarrow 0} \frac{u(x+\mathrm{i} t, y)+\mathrm{i} v(x+\mathrm{i} t, y)-(u(x, y)+\mathrm{i} v(x, y))}{\mathrm{i} t} \\
& =-\mathrm{i}\left(\lim _{t \rightarrow 0} \frac{u(x+t, y)-u(x, y)}{t}+\mathrm{i} \lim _{t \rightarrow 0} \frac{v(x+t, y)-v(x, y)}{t}\right) \\
& =-\mathrm{i} \frac{\partial u}{\partial x}+\frac{\partial v}{\partial x}
\end{aligned}
$$

Since both limits are equal, as seen in Eq. (25.1), we obtain what we call the CauchyRiemann equation for holomorphic functions:

Theorem 25.0.2 (Cauchy-Riemann). Let $D \subseteq \mathbf{C}$ be an open set, and $f=u+\mathrm{i} v: D \rightarrow \mathbf{C}$ be any function - where $u, v: \mathbf{R}^{2} \rightrightarrows \mathbf{R}$ are differentiable functions. The function $f$ is holomorphic if and only if the following two equations are satisfied:

$$
\frac{\partial u}{\partial x}=\frac{\partial v}{\partial y}, \quad \text { and } \quad \frac{\partial v}{\partial x}=-\frac{\partial u}{\partial y} .
$$

Corollary 25.0.3. The complex derivative of the holomorphic function $f$ satisfies the following equalities:

$$
f^{\prime}(z)=\frac{\partial u}{\partial x}+\mathrm{i} \frac{\partial v}{\partial x}=\frac{\partial v}{\partial y}-\mathrm{i} \frac{\partial u}{\partial y}
$$

Corollary 25.0.4. The squared absolute value of the first order derivative of $f$ is equal to the Jacobian of the $\mathbf{R}$-vector field $F: D \rightarrow \mathbf{R}^{2}$ given by $(x, y) \mapsto(u(x, y), v(x, y))$ :

$$
\begin{aligned}
\left|f^{\prime}(z)\right|^{2} & =\operatorname{Jac} F(z) \\
& =\frac{\partial u}{\partial x} \frac{\partial v}{\partial y}-\frac{\partial u}{\partial y} \frac{\partial v}{\partial x} \\
& =\left(\frac{\partial u}{\partial x}\right)^{2}+\left(\frac{\partial v}{\partial x}\right)^{2} \\
& =\left(\frac{\partial v}{\partial y}\right)^{2}+\left(\frac{\partial u}{\partial y}\right)^{2}
\end{aligned}
$$

### 25.1 Angles \& Holomorphic Maps

Let $D \subseteq \mathbf{C}$ be an open set and let $\gamma=x+\mathrm{i} y:[a, b] \rightarrow D$ be a differentiable curve parametrized by the real valued interval $[a, b] \subseteq \mathbf{R}$. Consider a holomorphic function $f: D \rightarrow \mathbf{C}$.

Let's take a closer look at the composition $f \gamma:[a, b] \rightarrow \mathbf{C}$. We can interpret $\gamma^{\prime}(t)$ as defining a direction at the tangent space $T_{\gamma(t)} D$ whenever $\gamma^{\prime}(t) \neq 0$. Consider $z_{0}:=\gamma(t)$ and let $\eta:[a, b] \rightarrow D$ be another differentiable curve such that $\eta(h)=z_{0}$ for some $h \in[a, b]$.

Theorem 25.1.1 (Holomorphic functions are conformal). If $f^{\prime}\left(z_{0}\right) \neq 0$ then the angle between the directional vectors defined by $\gamma^{\prime}(t)$ and $\eta^{\prime}(h)$ at $T_{z_{0}} D$ is the same as the angle between the directional vectors defined by $(f \gamma)^{\prime}(t)$ and $(f \eta)^{\prime}(h)$.

Proof. Define an inner product $\langle-,-\rangle: \mathbf{C} \times \mathbf{C} \rightarrow \mathbf{R}$ given by $\langle z, w\rangle:=\operatorname{Re}(z \bar{w})$. If $z=a+\mathrm{i} b$ and $w=c+\mathrm{i} d$ then

$$
\langle z, w\rangle=\operatorname{Re}(a c+b d+\mathrm{i}(b c-a d))=a c+b d=\langle(a, b),(c, d)\rangle_{\mathbf{R}^{2}}
$$

where $\langle-,-\rangle_{\mathbf{R}^{2}}$ is the canonical inner product of the real vector space $\mathbf{R}^{2}$.
Let $\theta(z, w)$ be the angle between $z$ and $w$, then

$$
\cos (\theta(z, w))=\frac{\langle z, w\rangle}{|z||w|}
$$

and since $\sin (t)=\cos (t-\pi / 2)$ it follows that

$$
\sin (\theta(z, w))=\frac{\langle z,-\mathrm{i} w\rangle}{|z||w|}
$$

continue proof

## Chapter 26

## Normed Vector Spaces

Remark 26.0.1. Throughout this whole chapter we shall denote by $k$ a field that is either $\mathbf{C}$ or $\mathbf{R}$ - we'll adopt the use of $|-|: k \rightarrow \mathbf{R}_{\geqslant 0}$ for the standard norm of the underlying field in order to distinguish that from the norm of the vector spaces.

### 26.1 Norms on Spaces

Definition 26.1.1 ((Pre)Norm and (pre)normed vector spaces). Let $E$ be a $k$-vector space. We say that a map $\|-\|: E \rightarrow \mathbf{R}_{\geqslant 0}$ is a pre-norm in $E$ if for all $x, y \in E$ and $\lambda \in k$ the map satisfies the following properties
(a) Product by scalar: $\|\lambda x\|=|\lambda|\|x\|$.
(b) Triangle inequality: $\|x+y\| \leqslant\|x\|+\|y\|$.

Moreover, if || $-\|$ satisfies the following additional condition, it is called a norm.
(c) If $\|x\|=0$ then $x=0$.

The vector space $E$ endowed with the (pre)norm $\|-\|$ is called a (pre)normed vector space.

It should be noted immediately that the first condition for a pre-norm implies in $\|0\|=0$, hence the last condition for $\|-\|$ to be a norm can be substituted equivalently by " $\|x\|=0$ if and only if $x=0$ ". Moreover, in a prenormed space, it's not possible to assert that a sequence has a unique limit, property which is only ensured by the last condition.

A simple property that can be extracted from Definition 26.1.1, is that any subspace of a (pre)normed $k$-vector space is itself (pre)normed with the naturally inherited norm.

Example 26.1.2 (Norms from maps). Let $E$ be a $k$-vector space and let's consider any functional $f \in E^{*}$. We can build a pre-norm from the linear map $f$ by defining a map $\|-\| f: E \rightarrow \mathbf{R}_{\geqslant 0}$ given by $\|x\|_{f}:=|f(x)|$ - which clearly satisfies all of the required conditions for a norm. Note however that for $\|-\|_{f}$ to satisfy the last condition, it must be true that $\operatorname{ker} f=0$, that is, $\operatorname{dim}_{k} E=1$ necessarily for $\|-\|_{f}$ to be a norm.

Example 26.1.3 (Metrics from norms). Given a (pre)normed $k$-vector space $(E,\|-\|$ ), we can naturally define a metric in $E$ to be a map $d: E \times E \rightarrow \mathbf{R}_{\geqslant 0}$ given by $d(x, y):=\|x-y\|$ — which makes $E$ into a (pre)metric space (refer to Definition 12.3.1). We now verify each of the conditions for a pre-metric:
(a) If $x, y \in E$ are any elements, then

$$
d(x, y)=\|x-y\|=|-1|\|x-y\|=\|y-x\|=d(y, x)
$$

(b) If $z \in E$ is another element, then

$$
d(x, z)=\|x-z\| \leqslant\|x-y\|+\|y-z\|=d(x, y)+d(y, z)
$$

Moreover, if $\|-\|$ is a norm, then we are also able to satisfy:
(c) Given $x, y \in E$ such that $d(x, y)=0$, then $\|x-y\|=0$ and since $\|-\|$ is a norm, we obtain $x-y=0$, thus $x=y$ as wanted.

Proposition 26.1.4. Let $(E,\|-\|)$ be a (pre)normed $k$-vector space, and $d: E \times E \rightarrow \mathbf{R}_{\geqslant 0}$ be the (pre)metric induced by $\|-\|$. Then, $E$ is a Hausdorff space if and only if $\|-\|$ is a norm.

Proof. Suppose $E$ is a Hausdorff space and let $x, y \in E$ be such that $\|x-y\|=0$, then in particular $d(x, y)=0$ and therefore any open ball $B_{x}(r)$ centred at $x$, for $r>0$, does also contain $y$ - but since $E$ is Hausdorff, this can only be true if $x=y$, thus $x-y=0$ and we obtain that $\|-\|$ is a norm, and $d$ is a metric.

Now, assume that $\|-\|$ is a norm, thus $d$ is a metric and since any metric space is Hausdorff (see Proposition 12.4.9), in particular, E is Hausdorff.

Lemma 26.1.5. Let $(E,\|-\|)$ be a normed $k$-vector space, then the following is true for any given $x, y, z \in E$ :
(a) If $d: E \times E \rightarrow \mathbf{R}_{\geqslant 0}$ is the metric induced by $\|-\|$, then $d(x+y, y+z)=d(x, z)$.
(b) The inequality $|\|x\|-\|y\|| \leqslant\|x-y\|$ holds.

Proof. (a) $d(x+y, y+z)=\|(x+y)-(y+z)\|=\|x-z\|=d(x, z)$.
(b) Notice that $\|x\|=\|(x-y)+y\| \leqslant\|x-y\|+\|y\|$, therefore $\|x\|-\|y\| \leqslant\|x-y\|$. Moreover, symmetrically we have that $\|y\|-\|x\| \leqslant\|y-x\|=\|x-y\|$, therefore,

$$
-\|x-y\| \leqslant\|x\|-\|y\| \leqslant\|x-y\|
$$

as wanted.

Proposition 26.1.6. Any norm in an $\mathbf{R}$-vector space is continuous.
Proof. This follows directly from the inequality obtained in item (b) of Lemma 26.1.5. $\quad$,

Proposition 26.1.7. Given a (pre)normed $k$-vector space $E$, if we endow the products $E \times E$ and $k \times E$ with the product topology, addition of vectors and multiplication of vectors by scalars are both continuous maps.

Proof. Given sequences $x_{j} \rightarrow x$ and $y_{j} \rightarrow y$ for any $x, y \in E$, then $x_{j}+y_{j} \rightarrow x+y$, let $\varepsilon>0$ be any bound, then choose a common $n \in \mathbf{N}$ for which $\left\|x-x_{j}\right\|<\frac{\varepsilon}{2}$ and $\left\|y-y_{j}\right\|<\frac{\varepsilon}{2}$ for all $j \geqslant n$. Therefore, by the triangle inequality

$$
\left\|(x+y)-\left(x_{j}-y_{j}\right)\right\| \leqslant\left\|x-x_{j}\right\|-\left\|y-y_{j}\right\|<\varepsilon \text { for all } j>n
$$

where we thus conclude that $x_{j}+y_{j} \rightarrow x+y-$ all limits are well defined since $E$ is Hausdorff.

Let $\lambda_{j} \rightarrow \lambda$ be a convergent sequence in $k$ for any given $\lambda \in k$. Since convergent sequences are bounded, let $M \in k$ be a bound for $\left(x_{j}\right)_{j \in \mathbf{N}}$, that is, $\left|x_{j}\right|<M$ - also, define $M^{\prime}:=\max (M,|\lambda|)$. Given any $\varepsilon>0$ we choose a common $n \in \mathbf{N}$ for which $\left|\lambda-\lambda_{j}\right|<\frac{\varepsilon}{2 M^{\prime}}$ and $\left\|x-x_{j}\right\|<\frac{\varepsilon}{2 M^{\prime}}$ for all $j \geqslant n$. With this we can use the triangle inequality to find

$$
\begin{aligned}
\left\|\lambda x-\lambda_{j} x_{j}\right\| & =\left\|\left(\lambda-\lambda_{j}\right) x_{j}+\lambda\left(x-x_{j}\right)\right\| \\
& \leqslant\left\|\left(\lambda-\lambda_{j}\right) x_{j}\right\|+\left\|\lambda\left(x-x_{j}\right)\right\| \\
& =\left|\lambda-\lambda_{j}\right|\left\|x_{j}\right\|+|\lambda|\left\|x-x_{j}\right\| \\
& =\frac{\varepsilon}{2 M^{\prime}}\left\|x_{j}\right\|+|\lambda| \frac{\varepsilon}{2 M^{\prime}} \\
& <\frac{\varepsilon}{2 M^{\prime}} M^{\prime}+M^{\prime} \frac{\varepsilon}{2 M^{\prime}} \\
& =\varepsilon
\end{aligned}
$$

that is, $\lambda_{j} x_{j} \rightarrow \lambda x$ as wanted.
Definition 26.1.8 (Equivalence of norms). Given vector space $E$ and two norms, $\|-\|_{1}$ and $\|-\|_{2}$, on $E$, we say that such norms are equivalent if there exists $a, b>0$ such that for all $v \in E$ we have the inequalities

$$
a\|v\|_{1} \leqslant\|v\|_{2} \leqslant b\|v\|_{1} .
$$

Proposition 26.1.9. Two norms are equivalent if and only if they induce the same topology.

Proof. Let $E$ be a $k$-vector space and both $\|-\|_{1}$ and $\|-\|_{2}$ be norms on $E$. Let $U \subseteq\left(E,\|-\|_{1}\right)$ be open and, for any given $x \in U$, let $B_{x}(r) \subseteq\left(E,\|-\|_{1}\right)$ be an open ball centred at $x$. Notice that if $b>0$ is the right constant in Definition 26.1.8, then by choosing the open ball $B_{x}(r / b) \subseteq\left(E,\|-\|_{2}\right)$, we are able to obtain that for all $y \in B_{x}(r / b)$ we have

$$
\|x-y\|_{1} \leqslant b\|x-y\|_{2}<r
$$

thus $y \in B_{x}(r)$ in $\|-\|_{1}$ - this implies in $B_{x}(c r) \subseteq B_{x}(r)$ in $\|-\|_{2}$ and therefore we conclude that $B_{x}(r)$ is open in $\left(E,\|-\|_{2}\right)$. Notice that the proof can be mirrored for the other case, thus we are done.

## Quotient Space

Example 26.1.10 (Pre-norm on quotients of normed spaces). Let $(E,\|-\|)$ be a pre-normed $k$-vector space and $F \subseteq E$ a subspace. We can define a pre-norm $\|-\| \sim: E / F \rightarrow \mathbf{R}_{\geqslant 0}$ on the quotient space $E / F$ as

$$
\|[v]\|_{\sim}:=\inf _{u \in[v]}\|u\|,
$$

that is, the infimum of the norm of the representatives of the class. This indeed defines a pre-norm since, if $[v],[w] \in E / F$, then

$$
\begin{align*}
\|[v]+[w]\|_{\sim} & =\inf \{\|x+y\|: x+y \in[v]+[w]\} \\
& \leqslant \inf \{\|x\|+\|y\|: x \in[v] \text { and } y \in[w]\} \tag{26.1}
\end{align*}
$$

We claim that $\inf \{\|x\|+\|y\|\} \leqslant \inf \|x\|+\inf \|y\|$. To prove that, assume that both $\inf \|x\|$ and $\inf \|y\|$ are finite - otherwise, if one of them is infinite, the inequality is trivially true. For any $\varepsilon>0$ we can find $x^{\prime} \in[v]$ and $y^{\prime} \in[w]$ for which

$$
\begin{aligned}
& \inf _{x \in[v]}\|x\| \leqslant\left\|x^{\prime}\right\| \leqslant \inf _{x \in[v]}\|x\|+\varepsilon, \\
& \inf _{y \in[v]}\|y\| \leqslant\left\|y^{\prime}\right\| \leqslant \inf _{y \in[v]}\|y\|+\varepsilon .
\end{aligned}
$$

Therefore we find that $\inf \{\|x\|+\|y\|\} \leqslant\left\|x^{\prime}\right\|+\left\|y^{\prime}\right\|$ therefore $\inf \{\|x\|+\|y\|\} \leqslant \inf \|x\|+$ $\inf \|y\|+2 \varepsilon$, but since $\varepsilon$ may be indefinitely little, we find that indeed $\inf \{\|x\|+\|y\|\} \leqslant$ $\inf \|x\|+\inf \|y\|$. Hence, we can turn to Eq. (26.1) and conclude that

$$
\begin{aligned}
\|[v]+[w]\|_{\sim} & \leqslant \inf \{\|x\|+\|y\|: x \in[v] \text { and } y \in[w]\} \\
& \leqslant \inf _{x \in[v]}\|x\|+\inf _{y \in[w]}\|y\| \\
& =\|[v]\|_{\sim}+\|[w]\|_{\sim},
\end{aligned}
$$

thus satisfying the triangle inequality. The condition that $\|\lambda v\|_{\sim}=|\lambda|\|v\|_{\sim}$ is trivially obtained - hence $\|-\|_{\sim}$ is a pre-norm in $E / F$.

Proposition 26.1.11 (Quotient norm). The pre-norm $\|-\|_{\sim}$ described in Example 26.1.10 is a norm if and only if the subspace $F$ is closed in $E$.

Proof. Let $[v] \in E / F$ be any class such that $\|[v]\| \|_{\sim}=0$, therefore, we must be able to find a sequence $\left(v_{j}\right)_{j \in \mathbf{N}}$ of elements $v_{j} \in[v]$ such that $\left\|v_{j}\right\| \rightarrow 0$ as $j \rightarrow \infty$. Moreover, choosing $v \in[v]$ to be any representative, $\operatorname{since} v_{j}-v \in F$ and $F$ is closed normed space, the convergence of the norm to zero implies that $v_{j} \rightarrow 0$ for $j \rightarrow \infty$ in $F$ - therefore, we conclude that $F \ni v_{j}-v \rightarrow-v$ and thus $-v \in F$ from the closeness property, in particular, we find that $v \in F$.

We now claim that since $v \in F$, then $[v]=[0]$. Let $u \in[v]$ be any representative, then $u-v \in F$ and since $v \in F$ by assumption, then in particular $u \in F$, hence $u \in[0]$ - that is, $[v] \subseteq[0]$. Now, if $w \in[0]$, from definition we obtain $w \in F$, but since $v \in F$ then in particular $w-v \in F$ hence $w \in[v]$ - which implies that $[0] \subseteq[v]$ and thus $[v]=[0]$ for all $v \in F$. Therefore, from this claim we obtain that $\|[v]\| \sim=0$ implies $[v]=[0]=F$, the zero element of $E / F$.

Proposition 26.1.12 (Norm out of pre-norm). Let $(E,\|-\|)$ be a pre-normed $k$-vector space and $E_{0}:=\{x \in E:\|x\|=0\}$. Then, the map $\|-\| \sim: E / E_{0} \rightarrow \mathbf{R} \geqslant 0$ defined by $\|[x]\|_{\sim}:=\|x\|$ is well defined and is a norm for the space $E / E_{0}$.

Proof. Let $[x] \in E / E_{0}$ be a class such that $\|[x]\| \sim=0$. Choose any representative $x \in[x]$ and notice that since $\|x\|=0$ then $x \in E_{0}$, that is, $[x] \subseteq E_{0}$ - moreover, if $y \in E_{0}$, then surely $y \in[x]$, that is, $E_{0} \subseteq[x]$. This shows that $[x]=E_{0}$, where $E_{0}=[0] \in E / E_{0}$.

## Examples of Normed Spaces

The following is an immediate proposition, so we won't bother to write down the proof.

Lemma 26.1.13 (Complex conjugate space norm). Let $(E,\|-\|)$ be a pre-normed $\mathbf{C}$ vector space. Then the naturally induced map $\|-\|: \bar{E} \rightarrow \mathbf{R}_{\geqslant 0}$ is a pre-norm for the complex conjugate space $\bar{E}$.

Proposition 26.1.14. Let $E$ be a pre-normed $k$-vector space, $F \subseteq E$ be a closed subspace, and $x \in E \backslash F$. Then there exists a scalar $C>0$ for which, for every given scalar $\lambda \in k$ and vector $y \in F$, we have

$$
|\lambda| \leqslant C\|\lambda x+y\|
$$

Proof. If $\lambda=0$ then the proposition follows trivially. Otherwise, let $\lambda \neq 0$ and notice that since $F$ is closed, there must exist $\theta>0$ for which $\|x-y\| \geqslant \theta$ for every given $y \in F$ - since $x$ lies outside of $F$. In particular, since $-\frac{1}{\lambda} y \in F$, then $\left\|x-\left(-\frac{1}{\lambda} y\right)\right\| \geqslant \theta$ - notice that such choice of vector was not made arbitrarily since

$$
\|\lambda x+y\|=|\lambda|\|x-(-y / \lambda)\| \geqslant|\lambda| \theta
$$

therefore, if we choose $C:=\frac{1}{\theta}$, we obtain the desired inequality.
Example 26.1.15 ( $p$-norms). The following are recurrent norms on two of the most relevant spaces to our analytical study of normed $k$-vector spaces:
(a) For every integer $1 \leqslant p<\infty$ we define a norm $\|-\|_{p}: k^{n} \rightarrow \mathbf{R}_{\geqslant 0}$ defined by, for all $x \in k^{n}$,

$$
\|x\|_{p}:=\left(\sum_{j=1}^{n}\left|x_{j}\right|^{p}\right)^{1 / p}
$$

(b) The infinite case for the space $k^{n}$ is defined by a map $\|-\|_{\infty}: k^{n} \rightarrow \mathbf{R}_{\geqslant 0}$ given by

$$
\|x\|_{\infty}:=\max _{1 \leqslant j \leqslant n}\left|x_{j}\right| .
$$

(c) Let $\ell^{p}(J) \subseteq k^{J}$ be the $k$-vector subspace consisting of maps $f: J \rightarrow k$ with countable support - that is, $\sum_{t \in J}|f(t)|^{p}<\infty$. For each integer $1 \leqslant p<\infty$, we define a norm $\|-\|_{p}: \ell^{p}(J) \rightarrow \mathbf{R}_{\geqslant 0}$ given by

$$
\|f\|_{p}:=\left(\sum_{t \in J}|f(t)|^{p}\right)^{1 / p}
$$

(d) For the infinite case, we define $\ell^{\infty}(J) \subseteq k^{J}$ to be the $k$-vector subspace consisting of maps $f: J \rightarrow k$ such that $\sup _{t \in J}|f(t)|<\infty$. We define the norm $\|-\|_{\infty}: \ell^{\infty}(J) \rightarrow \mathbf{R} \geqslant 0$ by

$$
\|f\|_{\infty}:=\sup _{t \in J}|f(t)|
$$

In fact, each one of the above spaces is Banach, but we'll prove this later. We need now to prove that these are indeed normed vector spaces.
Proof. (a) Let $x \in k^{n}$ be any vector and $\lambda \in k$ any scalar, then

$$
\|\lambda x\|_{p}=\left(\sum_{j}\left|\lambda x_{j}\right|^{p}\right)^{1 / p}=\left(|\lambda|^{p} \sum_{j}\left|x_{j}\right|^{p}\right)^{1 / p}=|\lambda|\|x\|_{p}
$$

Moreover, if $y \in k^{n}$ is another vector, then since $p \geqslant 1$ we may just use Minkowski's inequalities (see Proposition A.1.9) in order to obtain the triangle inequality - it should be noted that $p<1$ does not yield a valid triangle inequality, hence justifying the restriction $p \geqslant 1$. Also, if $\|z\|=0$ for some $z \in k^{n}$, then $z_{j}=0$ for all $1 \leqslant j \leqslant n$ - since $\left|z_{j}\right|=0-$ and thus $z=0$.
(b) The map $\|-\|_{\infty}: k^{n} \rightarrow \mathbf{R}_{\geqslant 0}$ is clearly a norm.
(c)
(d) Let $\lambda \in k$ and $f, g \in \ell^{\infty}(J)$ be any elements. Since $\sup _{t}|f(t)|<\infty$, we conclude that $\|\lambda f\|=\sup _{t}|\lambda f(t)|=\sup _{t}|\lambda||f(t)|=|\lambda| \sup _{t}|f(t)|$, thus $\|\lambda f\|=|\lambda|\|f\|$. Moreover, we have $\sup _{t}|f(t)+g(t)| \leqslant \sup _{t}(|f(t)|+|g(t)|) \leqslant \sup _{t}|f(t)|+\sup _{t}|g(t)|<\infty$, therefore $\|f+g\| \leqslant\|f\|+\|g\|$. Also, if $h \in \ell^{\infty}(J)$ is a map such that $\|h\|=0$, then $\sup _{t}|h(t)|=0$, which by the definition of the supremum of a set and since $|h(t)| \geqslant 0$, we conclude that $h(t)=0$ for every $t \in J$ - that is, $h=0$.
Prove that the maps define norms for $\ell^{p}$ and $\ell^{\infty}$, moreover, prove that each of the mentioned spaces is Banach.

Lemma 26.1.16. Let $\|-\|_{p},\|-\|_{\infty}: k^{n} \rightrightarrows \mathbf{R}_{\geqslant 0}$ be the norms defined in Example 26.1.15, then the following inequality holds for all $1 \leqslant p<\infty$ and all $x \in k^{n}$ :

$$
\|x\|_{\infty} \leqslant\|x\|_{p} \leqslant n^{1 / p}\|x\|_{\infty}
$$

Therefore $\|-\|_{p}$ and $\|-\|_{\infty}$ are equivalent norms.
Proof. Since $\left(\sum_{j}\left|x_{j}\right|^{p}\right)^{1 / p} \leqslant\left(\sum_{j}\left(\max _{j}\left|x_{j}\right|\right)^{p}\right)^{1 / p}=n^{1 / p} \max _{j}\left|x_{j}\right|$, we obtain $\|x\|_{p} \leqslant$ $n^{1 / p}\|x\|_{\infty}$. Moreover, clearly $\|x\|_{\infty} \leqslant\|x\|_{p}$, thus the proposition follows.

■
Example 26.1.17 ( $\ell^{\infty}$ space of maps). Let $X$ be any set and define $\ell^{\infty}(X)$ to be a $k$-vector subspace of $k^{X}$, consisting of all bounded maps. We define a norm $\|-\|_{\infty}: \ell^{\infty}(X) \rightarrow \mathbf{R} \geqslant 0$ to be the map given by

$$
\|f\|_{\infty}:=\sup _{t \in X}|f(t)| .
$$

Such norm is called the uniform norm on the function space.

Example 26.1.18 (Uniform convergence norm). Let $X$ be a Hausdorff space and $\Omega \subseteq X$ be a compact set. If $C(\Omega, k)$ denotes the $k$-vector space of continuous functionals $\Omega \rightarrow k$, the map $\|-\|_{\infty}: C(\Omega, k) \rightarrow \mathbf{R}_{\geqslant 0}$ given by

$$
\|f\|_{\infty}:=\sup _{t \in \Omega}|f(t)|=\max _{t \in \Omega}|f(t)|
$$

defines a norm in $C(\Omega, k)$. Moreover, the normed $k$-vector space $\left(C(\Omega, k),\|-\|_{\infty}\right)$ is Banach.

Proof.
Prove

### 26.2 Properties of Normed Vector Spaces

## Finite Dimensional

Lemma 26.2.1. Let $E$ be an $n$-dimensional normed $k$-vector space and $\left\{x_{1}, \ldots, x_{n}\right\}$ be a basis of $E$. There exists a scalar $C>0$ such that, for every choice of scalars $\left\{\lambda_{1}, \ldots, \lambda_{n}\right\} \subseteq k$, we have

$$
\left\|\sum_{j=1}^{n} \lambda_{j} x_{j}\right\| \geqslant C \sum_{j=1}^{n}\left|\lambda_{j}\right| .
$$

Proof. For the sake of brevity, denote $S:=\sum_{j=1}^{n}\left|\lambda_{j}\right|$. If $S=0$ then the lemma follows trivially. Suppose that $S>0$ and, for the sake of contradiction, that there exists no scalar $C>0$ such that $\left\|\sum_{j}\left|\lambda_{j} x_{j}\right|\right\| \geqslant C \sum_{j}\left|\lambda_{j}\right|$ - this implies that for all integer $p \geqslant 1$ there exists a point $y_{p}:=\sum_{j=1}^{n} \lambda_{j}(p) x_{j}$ such that

$$
\begin{equation*}
\sum_{j=1}^{n}\left|\lambda_{j}(p)\right| \geqslant p\left\|y_{p}\right\| \tag{26.2}
\end{equation*}
$$

Moreover, for each $p \geqslant 1$, we may as well construct $z_{p}:=\sum_{j=1}^{n} \alpha_{j}(p) x_{j}$ - where $\alpha_{j}: \mathbf{N} \rightarrow k$ for all $1 \leqslant j \leqslant n$ - for which

$$
\alpha_{j}(p):=\frac{\lambda_{j}(p)}{\sum_{j=1}^{n}\left|\lambda_{j}(p)\right|}
$$

This ensures us that $\sum_{j} \alpha_{j}(p)=1$ - hence $\left|\alpha_{j}(p)\right| \leqslant 1$. Dividing Eq. (26.2) by $\sum_{j}\left|\lambda_{j}(p)\right|$ we find

$$
\begin{equation*}
\frac{1}{p} \geqslant\left\|y_{p}\right\| \tag{26.3}
\end{equation*}
$$

For every fixed $1 \leqslant j \leqslant n$, the sequence $\left(\alpha_{j}(p)\right)_{p}$ is bounded and since we are working either with the complex or real numbers as the underlying field, we can conclude from Bolzano-Weirstraß theorem (see Theorem 14.3.2) that from the sequence
of bounded scalars one can extract a convergent subsequence $\left(\alpha_{j}\left(p^{\prime}\right)\right)_{p^{\prime}}$ - assume that $\alpha_{j}\left(p^{\prime}\right) \rightarrow \beta_{j}$ for some $\beta_{j} \in k$. This induces a subsequence of $\left(y_{p}\right)_{p}$, given by $\left(y_{p^{\prime}}\right)_{p^{\prime}}$, so that $y_{p^{\prime}} \rightarrow \sum_{j} \beta_{j} x_{j}-$ moreover, $\sum_{j}\left|\beta_{j}\right|=1$. However, from Eq. (26.3) we find that $y_{p} \rightarrow 0$, thus also $y_{p^{\prime}} \rightarrow 0$. Since $E$ is Hausdorff, the limit of the sequence $\left(y_{p^{\prime}}\right)_{p^{\prime}}$ must be unique, hence $\sum_{j} \beta_{j} x_{j}=0$ - which contradicts the initial hypothesis that the set $\left\{x_{1}, \ldots, x_{n}\right\}$ was linearly independent in $E$.

This shows one cannot build a sequence $\left(y_{p}\right)_{p}$ for which Eq. (26.2) is satisfied, hence the proposition follows.

Corollary 26.2.2. The unit ball is compact in a finite dimensional normed vector space.
Proof. Let $E$ be an $n$-dimensional $k$-vector space and $\left\{e_{1}, \ldots, e_{n}\right\}$ be a basis for $E$. Let $\left(x_{p}\right)_{p \in \mathbf{N}}$ be a sequence of points in the unit ball $B_{0}(1)$ (that is, $\left\|x_{p}\right\| \leqslant 1$ ) - we'll define for each $p \in \mathbf{N}$ that $x_{p}:=\sum_{j=1}^{n} \alpha_{j}(p) e_{j}$, where $\alpha_{j}: \mathbf{N} \rightarrow k$ for all $1 \leqslant j \leqslant n$. One concludes that, since $\left\|\alpha_{j}(p)\right\| \leqslant 1$ for each $1 \leqslant j \leqslant n$, we can use the BolzanoWeierstraß theorem (see Theorem 14.3.2) to conclude that we can extract a convergent subsequence $\left(\alpha_{j}\left(p^{\prime}\right)\right)_{p^{\prime}}$ from $\left(\alpha_{j}(p)\right)_{p \in \mathbf{N}}$ - for instance, assume that $\alpha_{j}(p) \rightarrow \beta_{j}$ for some $\beta_{j} \in k$. Then the induced subsequence $\left(x_{p^{\prime}}\right)_{p^{\prime}}$ is such that $x_{p^{\prime}} \rightarrow \sum_{j} \beta_{j} e_{j}:=x$. Moreover, for each $p^{\prime}$ we have the inequality $\|x\| \leqslant\left\|x-x_{p^{\prime}}\right\|+\left\|x_{p}\right\| \leqslant\left\|x-x_{p^{\prime}}\right\|+1$, thus as $p^{\prime} \rightarrow \infty$ we have $\left\|x-x_{p^{\prime}}\right\| \rightarrow 0$ and therefore $\|x\| \leqslant 1$ - that is, every sequence in $B_{0}(1)$ has a convergent subsequence in $B_{0}(1)$, which by Theorem 14.3.8 implies that $B_{0}(1)$ is compact in $E$.

Lemma 26.2.3. If $E$ is a finite dimensional $k$-vector space, then all norms in $E$ are equivalent.

Proof. Assume that $\left\{e_{1}, \ldots, e_{n}\right\}$ is a basis for $E$. We'll show that every norm is equivalent to $\|-\|: E \rightarrow \mathbf{R}_{\geqslant 0}$ given by $\left\|\sum_{j=1}^{n} \lambda_{j} e_{j}\right\|:=\sum_{j=1}^{n}\left|\lambda_{j}\right|$. Let $\|-\|^{\prime}: E \rightarrow \mathbf{R}_{\geqslant 0}$ be any other norm - then from the triangle inequality we have, for any point $x:=\sum_{j} \lambda_{j} e_{j}$ in E:

$$
\|x\|^{\prime} \leqslant \sum_{j=1}^{n}\left\|\lambda_{j} e_{j}\right\| \leqslant \max _{1 \leqslant j \leqslant n}\left\|e_{j}\right\| \sum_{j=1}^{n}\left|\lambda_{j}\right|=\max _{1 \leqslant j \leqslant n}\left\|e_{j}\right\|\|x\| .
$$

Defining $C:=\max _{j}\left\|e_{j}\right\|$, we have shown that $\|x\|^{\prime} \leqslant C\|x\|$.
For the last part, we must prove the existence of a scalar $B>0$ such that $\|x\| \leqslant B\|x\|^{\prime}$. We proceed by contradiction, that is, assuming we can choose a sequence of points $\left(x_{p}\right)_{p \in \mathbf{N}}$ such that

$$
\begin{equation*}
\left\|x_{p}\right\| \geqslant p\left\|x_{p}\right\|^{\prime} \tag{26.4}
\end{equation*}
$$

for all $p \in \mathbf{N}$. Let $\left(y_{p}\right)_{p \in \mathbf{N}}$ be the sequence defined by $y_{p}:=x_{p} /\left\|x_{p}\right\|$, so that $\left\|y_{p}\right\|=1$ - thus, by Eq. (26.4) we obtain the bound $\left\|y_{p}\right\|^{\prime} \leqslant 1 / p$, thus $y_{p} \rightarrow 0$. If we let $y_{p}:=\sum_{j} \alpha_{j}(p) e_{j}$, we obtain that $\sum_{j}\left|\alpha_{j}(p)\right|=1$ implies $\left|\alpha_{j}(p)\right| \leqslant 1$ - hence, for every fixed $1 \leqslant j \leqslant n$, the sequence $\left(\alpha_{j}(p)\right)_{p \in \mathbf{N}}$ contains a convergent subsequence $\left(\alpha_{j}\left(p^{\prime}\right)\right)_{p^{\prime}}$ such that $\alpha_{j}\left(p^{\prime}\right) \rightarrow \beta_{j}$ for some $\beta_{j} \in k$. Moreover, the limits are such that $\sum_{j}\left|\beta_{j}\right|=1$. Therefore, the induced subsequence $\left(y_{p^{\prime}}\right)_{p^{\prime}}$ is such that $y_{p^{\prime}} \rightarrow \sum_{j} \beta_{j} e_{j}$-however, since $E$ is Hausdorff, the sequence must have a unique limit, hence $\sum_{j} \beta_{j} e_{j}=0$, which is only possible if $\beta_{j}=0$ for all $1 \leqslant j \leqslant n$ - this contradicts the fact that $\sum_{j}\left|\beta_{j}\right|=1$. We
conclude that a sequence $\left(x_{p}\right)_{p \in \mathbf{N}}$ satisfying Eq. (26.4) must not exist, thus implying that there exists $B>0$ such that $\|x\| \leqslant B\|x\|^{\prime}$ for all $x \in E$.

Corollary 26.2.4. Every finite dimensional normed $k$-vector space is Banach.
Proof. We'll work on a $n$-dimensional $k$-vector space $(E,\|-\|)$ - where we let $\left\{e_{1}, \ldots, e_{n}\right\}$ be any basis and $\|-\|^{\prime}$ be the norm $\left\|\sum_{j} \lambda_{j} e_{j}\right\|^{\prime}:=\sum_{j}\left|\lambda_{j}\right|$. Let $\left(x_{p}\right)_{p \in \mathbf{N}}$ be any Cauchy sequence with respect to $\|-\|$, and define $x_{p}:=\sum_{j} \lambda_{j}(p) e_{j}$ for some map $\lambda_{j}: \mathbf{N} \rightarrow k$. From Lemma 26.2.3 we find that there exists $C>0$ for which $\|x\|^{\prime} \leqslant C\|x\|$ for every $x \in E$ - therefore $\left(x_{p}\right)_{p \in \mathbf{N}}$ is also Cauchy with respect to $\|-\|^{\prime}$. Hence, for all $\varepsilon>0$ there exists $N \in \mathbf{N}$ such that for all $p, q \geqslant N$ we have $\left\|x_{p}-x_{q}\right\|^{\prime}=\sum_{j}\left|\lambda_{j}(p)-\lambda_{j}(q)\right|<C \varepsilon$ - therefore for every $1 \leqslant j \leqslant n$ we have $\left|\lambda_{j}(p)-\lambda_{j}(q)\right|<C \varepsilon$, which implies that $\left(\lambda_{j}(p)\right)_{p \in \mathbf{N}}$ is Cauchy in $k$. Let for instance $\lambda_{j}(p) \rightarrow \alpha_{j}$ for some $\alpha_{j} \in k$. Moreover, notice that $\left\|x_{p}-\sum_{j} \alpha_{j} e_{j}\right\|^{\prime}=\sum_{j}\left|\lambda_{j}(p)-\alpha_{j}\right|$, which converges to zero as $p \rightarrow \infty-$ thus $x_{p} \rightarrow \sum_{j} \alpha_{j} e_{j}$, which proves that $\left(x_{p}\right)_{p \in \mathbf{N}}$ converges in $E$.

Proposition 26.2.5 (Compact sets). If $E$ is a finite dimensional normed $k$-vector space, a subset $\Omega \subseteq E$ is compact if and only if $\Omega$ is bounded and closed.

Proof. The first part comes from Proposition 14.2.20. For the second, let $\Omega$ be bounded and closed. If $\operatorname{dim} E=n$, we let $\left\{e_{1}, \ldots, e_{n}\right\}$ be a basis for $E$. Let $\left(x_{p}\right)_{p \in \mathbf{N}}$ be any sequence of points in $\Omega$ and assume that each $x_{p}$ has the form $x_{j}:=\sum_{i=1}^{n} \lambda_{i}(p) e_{i}$ for scalars $\lambda_{i}(p) \in k$. From the boundness of $\Omega$ one can find $M>0$ such that $\left\|x_{p}\right\| \leqslant C$ for every index $p \in \mathbf{N}$. Evoking Lemma 26.2.1 for each index $p \in \mathbf{N}$, we find that

$$
M \geqslant\left\|x_{p}\right\| \geqslant C \sum_{i=1}^{n}\left|\lambda_{i}(p)\right|
$$

for some $C>0$. Fixing any $1 \leqslant i \leqslant n$, one concludes that $\left|\lambda_{i}(p)\right| \leqslant M$ and therefore the sequence $\left(\lambda_{i}(p)\right)_{p \in \mathbf{N}}$ is bounded. Using Bolzano-Weierstraß theorem (see Theorem 14.3.2) there exists a convergent subsequence $\left(\lambda_{i}\left(p^{\prime}\right)\right)_{p^{\prime}}$ - assume for instance that $\lambda_{i}\left(p^{\prime}\right) \rightarrow \lambda_{i}$ for some $\lambda_{i} \in k$. Such subsequence induces another subsequence $\left(x_{p^{\prime}}\right)_{p^{\prime}}$, for which $x_{p^{\prime}} \rightarrow \sum_{i} \lambda_{i} e_{i}:=x$ - but since $\Omega$ is closed, we find $x \in \Omega$. Thus any sequence in $\Omega$ has a convergent subsequence in $\Omega$ - this shows that $\Omega$ is compact (see Theorem 14.3.8).

## $\ell^{\infty}(\mathbf{N})$ and $\ell^{p}(\mathbf{N})$ are Banach Spaces

Lemma 26.2.6. Every Cauchy sequence in a normed $k$-vector space is bounded.
Proof. Let $\left(x_{j}\right)_{j \in \mathbf{N}}$ be a Cauchy sequence in a normed $k$-vector space $E$. If $\varepsilon=1$, then there exists $N \in \mathbf{N}$ for which $i, j \geqslant N$ implies $\left\|x_{j}-x_{i}\right\|<1$ and therefore $\left\|x_{j}\right\| \leqslant\left\|x_{j}-x_{N}\right\|+\left\|x_{N}\right\|<1+\left\|x_{N}\right\|$. We conclude that the sequence is bounded:

$$
\left\|x_{j}\right\| \leqslant 1+\max _{0 \leqslant i \leqslant N}\left\|x_{i}\right\| .
$$

Proposition 26.2.7. The space $\ell^{\infty}(\mathbf{N})$ is Banach.
Proof. Let $x:=\left(x^{p}\right)_{p \in \mathbf{N}}$ denote any Cauchy sequence of points $x^{p} \in \ell^{\infty}(\mathbf{N})$ - that is, $x^{p}:=\left(x_{j}(p)\right)_{j \in \mathbf{N}}$ is itself a sequence - therefore for all $\varepsilon>0$ there exists $N \in \mathbf{N}$ such that for all $p, q \geqslant N$ we have

$$
\begin{equation*}
\left\|x^{p}-x^{q}\right\|=\sup _{j \in \mathbf{N}}\left|x_{j}(p)-x_{j}(q)\right|<\varepsilon \tag{26.5}
\end{equation*}
$$

We first construct a candidate for the limit of $x$. Let $j_{0} \in \mathbf{N}$ be any fixed index we'll show that the sequence $\left(x_{j_{0}}(p)\right)_{p \in \mathbf{N}}$ forms a Cauchy sequence in $k$. Notice that from Eq. (26.5) it is clear that $\sup _{j \in \mathbf{N}}\left|x_{j}(p)-x_{j}(q)\right|<\varepsilon$ implies in $\left|x_{j_{0}}(p)-x_{j_{0}}(q)\right|<\varepsilon$ for all $p, q>N$ - therefore $\left(x_{j_{0}}(p)\right)_{p \in \mathbf{N}}$ is indeed Cauchy, and since $k$ (either $\mathbf{R}$ or $\mathbf{C}$ ) is complete, there exists $y_{j_{0}} \in k$ for which $x_{j_{0}}(p) \rightarrow y_{j_{0}}$. Our candidate sequence will thus be formed by $y:=\left(y_{j}\right)_{j \in \mathbf{N}}$ - where each $y_{j}$ is constructed just as above. Indeed, $y \in \ell^{\infty}(\mathbf{N})$, since for every $j \in \mathbf{N}$ we have $\left|y_{j}\right|=\lim _{p \rightarrow \infty}\left|x_{j}(p)\right| \leqslant \lim _{p \rightarrow \infty}\left\|x^{p}\right\|_{\infty}$ and since $x$ is Cauchy, from Lemma 26.2.6, we find that there exists $C>0$ such that $\left|y_{j}\right| \leqslant\left\|x^{p}\right\|_{\infty}<C$ - hence $\|y\|_{\infty}=\sup _{j \in \mathbf{N}}\left|y_{j}\right|<\infty$.

We now show that $\left(y_{j}\right)_{j \in \mathbf{N}}$ is the limit of $\left(x^{p}\right)_{p \in \mathbf{N}}$. Let $\varepsilon>0$ be any bound and $N \in \mathbf{N}$ be such that $p, q>N$ implies $\left\|x^{p}-x^{q}\right\|_{\infty}<\varepsilon$ - that is, for any fixed $j_{0} \in \mathbf{N}$ we have $\left|x_{j_{0}}^{p}-x_{j_{0}}^{q}\right|<\varepsilon$. Moreover, if we let $q \rightarrow \infty$ we'll find that $\left|x_{j_{0}}^{p}-x_{j_{0}}^{q}\right| \rightarrow\left|x_{j_{0}}^{p}-y_{j_{0}}\right|<\varepsilon-$ hence, since this must be true for any $j_{0} \in \mathbf{N}$, we obtain that $\|x\|_{\infty}=\sup _{j \in \mathbf{N}}\left|x_{j}^{q}-y_{j}\right|<\varepsilon$. This proves that $x_{j}(p) \rightarrow y_{j}$ as $p \rightarrow \infty$ and therefore $x^{p} \rightarrow y$ as wanted. Thus any Cauchy sequence converges and the limit is given by the sequence of the limit of the components.

Proposition 26.2.8. The space $\ell^{p}(\mathbf{N})$ is Banach for all $1 \leqslant p<\infty$.

## prove

## Common Properties Disregarding the Dimension of the Space

Proposition 26.2.9. Let $f: X \rightarrow Y$ be an $\mathbf{R}$-linear map between normed $\mathbf{R}$-vector spaces $X$ and $Y$. Then, $f$ is contiguous if and only if there exists a scalar $C>0$, called bound, for which $\|f(x)\|_{Y} \leqslant C\|x\|_{X}$ for all $x \in X$.

Proof. If $f$ is bounded by $C$, let $B$ be a basis for $X$ and consider any element $x:=$ $\sum_{v \in B} a_{v} v$. From linearity we have $f(x)=\sum_{v \in B} a_{v} f(v)$, thus

$$
\|f(x)\|_{Y}=\left\|\sum_{v \in B} a_{v} f(v)\right\|_{Y} \leqslant \sum_{v \in B}\left\|a_{v}\right\|_{\mathbf{R}}\|f(v)\|_{Y} \leqslant C\|x\|_{X} \sum_{v \in B}\left\|a_{v}\right\|_{\mathbf{R}}
$$

This boils down to $f=O\left(\mathrm{id}_{X}\right)$ - which implies in $f\left(x-x_{0}\right)=f(x)-f\left(x_{0}\right) \rightarrow 0$ as $x \rightarrow x_{0}$, where $x_{0} \in X$ is any point, that is, $f$ is continuous at any point of $X$. Even better than that, we can show that $f$ is uniformly continuous (I won't carry it out since it's equivalent to what we wrote in Proposition A.2.37).

For the opposite, suppose $f$ is continuous at 0 , then there will surely exist $\delta>0$ for which $\|x\|_{X}<\delta$ implies in $\|f(x)\|_{Y}<1$. Therefore, for any choice of non-zero $x \in X$, we find

$$
\left\|f\left(\frac{\delta}{\|x\|_{X}} x\right)\right\|_{Y}=\frac{\delta}{\|x\|_{X}}\|f(x)\|_{Y}<1
$$

therefore $\|f(x)\|<\frac{\|x\|_{X}}{\delta}$, thus $f$ is indeed bounded.
Proposition 26.2.10. Let $E, F$ and $G$ be normed vector spaces, and let $u: E \rightarrow F$ and $v: F \rightarrow G$ be continuous linear maps. Then, vu: $E \rightarrow G$ is also a continuous linear map, and

$$
\|v u\| \leqslant\|v\|\|u\|,
$$

where $\|f\|:=\sup _{x \in X}\|f(x)\|$ for a continuous linear map $f: X \rightarrow Y$ between normed vector spaces ${ }^{1}$.

Proof. The first assertion is trivial, since the composition of continuous maps is continuous, and the same is true for linear maps. Let $x \in E$ be any element, notice that

$$
\|v u(x)\|_{G} \leqslant\|v\|\|u(x)\|_{F} \leqslant\|v\|\|u\|\|x\|_{E},
$$

thus the inequality holds.
Proposition 26.2.11. A multilinear map $\phi: \prod_{j=1}^{n} E_{j} \rightarrow F$ between normed vector spaces $E_{1}, \ldots, E_{n}$, and $F$, is continuous if and only if there exists a bound $C>0$ such that, for every $x \in \prod_{j=1}^{n} E_{j}$,

$$
\|\phi(x)\|_{F} \leqslant C \prod_{j=1}^{n}\left\|x_{j}\right\|_{E_{j}} .
$$

Proposition 26.2.12. Let $E_{1}, \ldots, E_{r}$, and $F$ be normed vector spaces. There exists a canonical map from repeated continuous linear maps to the continuous multilinear maps, which is a continuous linear isomorphism, and is norm-preserving - that is, the canonical map

$$
\Phi: L\left(E_{1}, L\left(E_{2}, \ldots, L\left(E_{r}, F\right), \ldots\right)\right) \xrightarrow{\simeq} L^{n}\left(E_{1}, \ldots, E_{n} ; F\right)
$$

is a Banach isomorphism.
Proof. We define $\Phi$ by the following: if $\lambda \in L\left(E_{1}, L\left(E_{2}, \ldots, L\left(E_{n}, F\right) \ldots\right)\right)$ is given by

$$
\lambda\left(x_{1}\right)=\lambda_{2}, \text { where } \lambda_{2}\left(x_{2}\right)=\lambda_{3}, \ldots, \lambda_{n}\left(x_{n}\right)=y \in F
$$

we define $\Phi(\lambda):=\bar{\lambda} \in L\left(E_{1}, \ldots, E_{n} ; F\right)$ by the mapping

$$
\bar{\lambda}\left(x_{1}, \ldots, x_{n}\right):=\lambda\left(x_{1}\right)\left(x_{2}\right) \ldots\left(x_{n}\right),
$$

[^35]where $\lambda_{j}\left(x_{j}\right)\left(x_{j+1}\right) \ldots\left(x_{n}\right):=\lambda_{j-1}\left(x_{j-1}\right) \ldots\left(x_{n}\right)$ for every $1 \leqslant j \leqslant n-$ where $\lambda_{1}:=\lambda$.
Given $\lambda \in L\left(E_{1}, L\left(E_{2}, \ldots, L\left(E_{n}, F\right), \ldots\right)\right)$, the map $\bar{\lambda}$ is surely multilinear since each of the recursive arguments are linear. Moreover, notice that, for any $x \in \prod_{j=1}^{n} E_{j}$, we have
$$
\|\bar{\lambda}(x)\|_{F} \leqslant\left\|\lambda\left(x_{1}\right)\left(x_{2}\right) \ldots\left(x_{n}\right)\right\|_{F} \leqslant\|\lambda\| \prod_{j=1}^{n}\left\|x_{j}\right\|_{E_{j}},
$$
thus $\|\bar{\lambda}\| \leqslant\|\lambda\|$.
On the other hand, given $\bar{\phi} \in L\left(E_{1}, \ldots, E_{n} ; F\right)$, define the map $\phi=\Phi^{-1}(\bar{\phi})$ by
$$
\phi\left(x_{1}\right)\left(x_{2}\right) \ldots\left(x_{n}\right):=\bar{\phi}\left(x_{1}, \ldots, x_{n}\right) .
$$

Therefore

$$
\left\|\phi\left(x_{1}\right)\left(x_{2}\right) \ldots\left(x_{n}\right)\right\|_{F} \leqslant\|\bar{\phi}\| \prod_{j=1}^{n}\left\|x_{j}\right\|_{E_{j}}
$$

which shows that $\|\phi\| \leqslant\|\bar{\phi}\|$. We conclude that $\Phi(\lambda)=\bar{\lambda}$ for all repeating map $\lambda$.
Theorem 26.2.13 (Hahn-Banach). Let $E$ be a normed R-vector space, and $F \subseteq E$ be a subspace. Let $\lambda \in F^{*}$ be a functional with bound $C>0$. Then there exist an extension of $\lambda$ to a functional on $E$ with the same bound $C$ - that is, a map $\bar{\lambda}: E \rightarrow \mathbf{R}$ such that $\left.\bar{\lambda}\right|_{F}=\lambda$ and $\|\bar{\lambda}(x)\|_{\mathbf{R}} \leqslant C\|x\|_{E}$ for all $x \in E$.

Corollary 26.2.14 (Hahn-Banach). Let $E$ be a Banach space and $x \in E$ be a non-zero element. There exists a continuous linear map $\phi \in E^{*}$ such that $\phi(x) \neq 0$.

## Properties of Banach Spaces

Definition 26.2.15. We define a Banach isomorphism to be a continuous linear map $u: E \rightarrow F$, between Banach spaces $E$ and $F$, that is both invertible (there exists a continuous linear map $u^{-2}: F \rightarrow E$ that is the two-sided inverse of $u$ ), and norm preserving - that is, given any $x \in E$, we have $\|u(x)\|_{F}=\|x\|_{E}$. Banach isomorphisms may also be referenced to isometries in the literature.

Proposition 26.2.16 (Bijections are isomorphisms). Every continuous bijective R-linear map between topological vector spaces is an isomorphism.

Proposition 26.2.17 (Splitting). Let $E$ be a Banach space, and $F$ and $G$ be complementary closed subspaces of $E$ - that is, $E=F+G$ and $F \cap G=0$. Then the morphism $F \times G \rightarrow E$ given by $(f, g) \mapsto f+g$ is a continuous linear isomorphism.

### 26.3 Topological Vector Spaces

Definition 26.3.1. A topological vector space is a $k$-vector space together with a topology such that addition of vectors and the product by scalars are both continuous $k$-linear maps.

We denote by TVect $\mathbf{R}_{\mathbf{R}}$ the category consisting of topological $\mathbf{R}$-vector spaces together with morphisms, which are continuous $\mathbf{R}$-linear maps (which may also be referenced to by the term "top-linear").

Let $E$ be a topological vector space. The continuous $\mathbf{R}$-linear maps corresponding to the dual space $E^{*}=\operatorname{Mor}_{\text {TVect }_{\mathbf{R}}}(E, \mathbf{R})$, of a topological $\mathbf{R}$-vector space $E$, are called $\mathbf{R}$ forms. The collection of forms of the form $E \rightarrow \mathbf{R}$ will be conveniently separated in classes and denoted:

- $L(E)$ : the collection of continuous linear maps $E \rightarrow \mathbf{R}$.
- $L^{r}(E)$ : the collection of continuous $r$-multilinear maps $E^{r} \rightarrow \mathbf{R}$.
- $L_{\text {Sym }}^{r}(E)$ : the collection of continuous $r$-multilinear symmetric maps $E^{r} \rightarrow \mathbf{R}$.
- $L_{\text {Alt }}^{r}(E)$ : the collection of continuous $r$-multilinear alternating maps $E^{r} \rightarrow \mathbf{R}$.

Definition 26.3.2 (Locally convex). A topological vector space $E$ is said to be locally convex if, for every open set $U \subseteq E$, any pair of points $x, y \in U$ are such that $t x+(1-t) y \in$ $U$ for all $t \in[0,1]$.

Definition 26.3.3 (Banachable). A topological R-vector space $E$ is said to be banachable if $E$ is complete and its topology can be defined by a norm.

As a point of order, every time we mention a topological $\mathbf{R}$-vector space in the course of this chapter, we shall mean a banachable space.

Definition 26.3.4 (Norm of a morphism). Let $E$ and $F$ be topological R-vector spaces. In order to make $\operatorname{Mor}_{\mathrm{TVect}_{\mathbf{R}}}(E, F)$ into a topological $\mathbf{R}$-vector space, we can construct a norm for which, given a morphism $A: E \rightarrow F$, define $K:=\left\{k \in \mathbf{R}:\|A x\|_{F} \leqslant\right.$ $k\|x\|_{E}$, for all $\left.x \in E\right\}$, the norm of $A$ is

$$
\|A\|:=\sup _{k \in K} k
$$

If $\operatorname{Mor}_{\text {TVect }_{\mathbf{R}}}\left(E_{1}, \ldots, E_{n} ; F\right)$ is the collection of continuous R-multilinear maps, then we define similarly the norm of a continuous multilinear map $B: \prod_{j=1}^{n} E_{j} \rightarrow F$ as

$$
\|B\|:=\sup _{m \in M} m
$$

where $M:=\left\{m \in \mathbf{R}:\|B x\|_{F} \leqslant m \prod_{j=1}^{n}\left\|x_{j}\right\|_{E_{j}}\right.$, for all $\left.x \in E\right\}$.
Remark 26.3.5. From now on, $C^{p}$-morphism will refer to a map $f: U \rightarrow V$ between open subsets of Banach spaces such that $f$ is a continuous map of class $C^{p}$, where $p \leqslant \infty$.

## Part IX

## Machine Learning

## Chapter 27

## Introduction to Probability Theory \& Statistics

### 27.1 Probability

Definition 27.1.1 (Sample space \& events). We define the sample space of an experiment to be the set $\Omega$ composed of all possible outcomes.

An event is defined to be any subset $A \subseteq \Omega$ of the sample space. The event $A$ is said to have occurred whenever the outcome inhabits $A$. Two events $A, B \subseteq \Omega$ are said to be mutually exclusive whenever $A$ and $B$ are disjoint.

Definition 27.1.2 ( $\sigma$-algebra). Let $\Omega$ be a set. A family of subsets $\Sigma \subseteq 2^{\Omega}$ is called a $\sigma$-algebra if it satisfies the following conditions:
(a) The empty set is an element of $\Sigma$.
(b) If $A \in \Sigma$, then the complement $A^{\mathrm{c}}$ is an element of $\Sigma$.
(c) If $\left\{A_{j}\right\}_{j \in J}$ is a collection of elements of $\Sigma$ indexed by a countable set $J$, then the union $\bigcup_{j \in J} A_{j}$ is also contained in $\Sigma$.

From condition (b) it follows immediately that $\Omega \in \Sigma$.
Definition 27.1.3 (Probability function). Given a sample space $\Omega$ and an associated $\sigma$-algebra $\Sigma$, we define a probability function on $\Sigma$ to be a map $\mathbb{P}: \Sigma \rightarrow \mathbf{R}$ such that
(a) The map $\mathbb{P}$ is non-negative.
(b) The probability of the whole sample space is 1 , that is, $\mathbb{P}(\Sigma)=1$.
(c) Given a countable set of pairwise disjoint events $\left\{A_{j}\right\}_{j} \subseteq \Sigma$, then

$$
\mathbb{P}\left(\bigcup_{j \in J} A_{j}\right)=\sum_{j \in J} \mathbb{P}\left(A_{j}\right) .
$$

Lemma 27.1.4. Let $\Omega=\left\{s_{j}\right\}_{j \in J}$ be a countable sample space and $\Sigma$ be an associated $\sigma$-algebra. If $\mathbb{P}: \Sigma \rightarrow \mathbf{R}$ is a mapping associated with non-negative real numbers $\left\{p_{j}\right\}_{j \in J}$ with $\sum_{j \in J} p_{j}=1$, for which

$$
\mathbb{P}(A):=\sum_{j: s_{j} \in A} p_{j}
$$

for each $A \in \Sigma$, then $\mathbb{P}$ is a probability function on $\Sigma$.
Theorem 27.1.5. Let $\Sigma$ be a $\sigma$-algebra associated to a sample space $\Omega$, and $\mathbb{P}: \Sigma \rightarrow \mathbf{R}$ be a probability function. If $A, B \in \Sigma$ are any sets then the following is holds:
(a) $\mathbb{P}\left(A^{\mathrm{C}}\right)=1-\mathbb{P}(A)$.
(b) $\mathbb{P}(\varnothing)=0$.
(c) $\mathbb{P}(A) \leqslant 1$, therefore $\mathbb{P}(\Sigma) \subseteq[0,1]$.
(d) $\mathbb{P}\left(B \cap A^{\mathrm{c}}\right)=\mathbb{P}(B)-\mathbb{P}(A \cap B)$.
(e) $\mathbb{P}(A \cup B)=\mathbb{P}(A)+\mathbb{P}(B)-\mathbb{P}(A \cap B)$, hence $\mathbb{P}(A \cup B) \geqslant \mathbb{P}(A)+\mathbb{P}(B)-1$, which is known as the Bonferroni's inequality.
(f) If $A \subseteq B$ then $\mathbb{P}(A) \leqslant \mathbb{P}(B)$.

Moreover, if $\left\{C_{j}\right\}_{j \in J}$ is a countable partition of $\Sigma$, and $\left\{A_{i}\right\}_{i \in \mathbf{N}}$ is any family of elements of $\Sigma$, we also have
(g) $\mathbb{P}(A)=\sum_{j \in J} \mathbb{P}\left(A \cap C_{j}\right)$.
(h) $\mathbb{P}\left(\bigcup_{i \in \mathbf{N}} A_{i}\right) \leqslant \sum_{i \in \mathbf{N}} \mathbb{P}\left(A_{i}\right)$.

Proof. (a) Notice that $A \cup A^{\mathrm{c}}=\Sigma$, therefore $\mathbb{P}\left(A \cup A^{\mathrm{c}}\right)=\mathbb{P}(A)+\mathbb{P}\left(A^{\mathrm{c}}\right)=1$, which proves the statement.
(b) Since $\Sigma^{\mathrm{C}}=\varnothing$, then $\mathbb{P}(\varnothing)=1-\mathbb{P}(\Sigma)=0$.
(c) Since $\mathbb{P}$ is a non-negative map, then $\mathbb{P}\left(A^{\mathrm{c}}\right) \geqslant 0$, hence $\mathbb{P}(A)=1-\mathbb{P}\left(A^{\mathrm{c}}\right)$ implies in $\mathbb{P}(A) \leqslant 1$.
(d) Notice that in general $B=(B \cap A) \cup\left(B \cap A^{\mathrm{c}}\right)$, therefore $\mathbb{P}(B)=\mathbb{P}(B \cap A)+\mathbb{P}\left(B \cap A^{\mathrm{c}}\right)$, from which the formula follows.
(e) Since $A \cup B=A \cup\left(B \cap A^{\mathrm{c}}\right)$, and from the fact that the sets $A$ and $B \cap A^{\mathrm{c}}$ are disjoint, then $\mathbb{P}(A \cup B)=\mathbb{P}(A)+\mathbb{P}\left(B \cap A^{\mathrm{C}}\right)$. Using the result from the last item for $\mathbb{P}\left(B \cap A^{\mathrm{C}}\right)$ we obtain the required formula. For the inequality, it suffices to see that $\mathbb{P}(A \cap B) \leqslant 1$.
(f) Since $A \cap B=A$ then by the result of item (d) we obtain

$$
\mathbb{P}(A)=\mathbb{P}(A \cap B)=\mathbb{P}(B)-\mathbb{P}\left(B \cap A^{\mathrm{c}}\right)
$$

and since $\mathbb{P}\left(B \cap A^{\mathrm{C}}\right) \in[0,1]$, then $\mathbb{P}(A) \leqslant \mathbb{P}(B)$.
(g) Since $\bigcup_{j \in J} C_{j}=\Sigma$ we have

$$
A=A \cap \Sigma=A \cap\left(\bigcup_{j \in J} C_{j}\right)=\bigcup_{j \in J} A \cap C_{j},
$$

therefore $\mathbb{P}(A)=\mathbb{P}\left(\bigcup_{j \in J} A \cap C_{j}\right)=\sum_{j \in J} \mathbb{P}\left(A \cap C_{j}\right)$.
(h) We shall construct a collection $\left\{A_{i}^{\prime}\right\}_{i \in \mathbf{N}}$ of disjoint sets partitioning $\bigcup_{i \in \mathbf{N}} A_{i}$. To do so, define $A_{0}^{\prime}:=A_{0}$ and for any other $i \in \mathbf{N}_{>0}$ we take $A_{i}^{\prime}:=A_{i} \backslash \bigcup_{j=0}^{i-1} A_{j}$. To see that such collection is indeed a partition, let $i, j \in \mathbf{N}$ be any two distinct indices and notice that

$$
\begin{aligned}
A_{i}^{\prime} \cap A_{j}^{\prime} & =\left(A_{i} \backslash \bigcup_{k=0}^{i-1} A_{k}\right) \cap\left(A_{j} \backslash \bigcup_{k=0}^{j-1} A_{k}\right) \\
& =\left(A_{i} \cap\left(\bigcup_{k=0}^{i-1} A_{k}\right)^{\mathrm{c}}\right) \cap\left(A_{j} \cap\left(\bigcup_{k=0}^{j-1} A_{k}\right)^{\mathrm{c}}\right) \\
& =\left(A_{i} \cap\left(\bigcap_{k=0}^{i-1} A_{k}^{\mathrm{c}}\right)\right) \cap\left(A_{j} \cap\left(\bigcap_{k=0}^{j-1} A_{k}^{\mathrm{c}}\right)\right),
\end{aligned}
$$

from which, if $i>j$ then $A_{j}^{c} \subseteq A_{i} \cap \bigcap_{k=0}^{i-1} A_{k}$, thus the first term is disjoint from the second-the case for $i<j$ is symmetric. Since our new collection satisfies the pairwise disjoint condition, we find $\mathbb{P}\left(\bigcup_{i \in \mathbf{N}} A_{i}\right)=\sum_{i \in \mathbf{N}} \mathbb{P}\left(A_{i}^{\prime}\right)$. From our construction we know that $A_{i}^{\prime} \subseteq A_{i}$ thus $\mathbb{P}\left(A_{i}^{\prime}\right) \leqslant \mathbb{P}\left(A_{i}\right)$ and hence $\sum_{i \in \mathbf{N}} \mathbb{P}\left(A_{i}^{\prime}\right) \leqslant \sum_{i \in \mathbf{N}} \mathbb{P}\left(A_{i}\right)$, which proves the statement.

Definition 27.1.6 (Conditional probability). Let $A$ and $B$ be events in a sample space $\Omega$, with $\mathbb{P}(B)>0$-where $\mathbb{P}$ is a probability function. We define the conditional probability of $A$ given $B$ to be

$$
\mathbb{P}(A \mid B)=\frac{\mathbb{P}(A \cap B)}{\mathbb{P}(B)}
$$

In this case, the map $\mathbb{P}(-\mid B): \Sigma \rightarrow \mathbf{R}$ is a probability function associated to $\mathbb{P}$.
Moreover, by symmetry we have $\mathbb{P}(B \mid A)=\mathbb{P}(A \cap B) / \mathbb{P}(A)$, therefore one obtains

$$
\mathbb{P}(A \mid B)=\mathbb{P}(B \mid A) \frac{\mathbb{P}(A)}{\mathbb{P}(B)}
$$

When $A$ and $B$ are unrelated events-that is, $A \cap B=\varnothing$-one has both $\mathbb{P}(A \mid B)=$ $\mathbb{P}(B \mid A)=0$ since $\mathbb{P}(A \cap B)=0$.

Theorem 27.1.7 (Bayes' rule). Let $\left\{A_{j}\right\}_{j \in J}$ be a countable partition of the sample space, and let $B$ be any event. Then, for each $j \in J$ on has

$$
\mathbb{P}\left(A_{j} \mid B\right)=\frac{\mathbb{P}\left(B \mid A_{j}\right) \mathbb{P}\left(A_{j}\right)}{\sum_{j \in J} \mathbb{P}\left(B \mid A_{j}\right) \mathbb{P}\left(A_{j}\right)}
$$

Definition 27.1.8 (Statistically independent events). A pair of events $A$ and $B$ is said to be statistically independent from each other if it is the case that $\mathbb{P}(A \cap B)=\mathbb{P}(A) \mathbb{P}(B)$.

Theorem 27.1.9. If $A$ and $B$ are statistically independent events, then the following pairs of events are also independent:
(a) $A$ and $B^{c}$.
(b) $A^{\mathrm{c}}$ and $B$.
(c) $A^{\mathrm{c}}$ and $B^{\mathrm{c}}$.

Proof. (a) One has

$$
\begin{aligned}
\mathbb{P}\left(A \cap B^{\mathrm{C}}\right) & =\mathbb{P}(A)-\mathbb{P}(A \cap B) \\
& =\mathbb{P}(A)-\mathbb{P}(A) \mathbb{P}(B) \\
& =\mathbb{P}(A)(1-\mathbb{P}(B)) \\
& =\mathbb{P}(A) \mathbb{P}\left(B^{\mathrm{C}}\right) .
\end{aligned}
$$

(b) The symmetric argument can be applied to the case of $A^{\mathrm{c}}$ and $B$.
(c) For both complements, we have

$$
\begin{aligned}
\mathbb{P}\left(A^{\mathrm{c}} \cap B^{\mathrm{c}}\right) & =\mathbb{P}\left((A \cup B)^{\mathrm{c}}\right) \\
& =1-\mathbb{P}(A \cup B) \\
& =1-(\mathbb{P}(A)+\mathbb{P}(B)-\mathbb{P}(A \cap B)) \\
& =1-\mathbb{P}(A)-\mathbb{P}(B)+\mathbb{P}(A) \mathbb{P}(B) \\
& =(1-\mathbb{P}(A))(1-\mathbb{P}(B)) \\
& =\mathbb{P}\left(A^{\mathrm{c}}\right) \mathbb{P}\left(B^{\mathrm{c}}\right) .
\end{aligned}
$$

Remark 27.1.10. Let $A, B$ and $C$ be three events in our sample space such that

$$
\mathbb{P}(A \cap B \cap C)=\mathbb{P}(A) \mathbb{P}(B) \mathbb{P}(C)
$$

It is not necessarily true that $A, B$ and $C$ are pairwise disjoint-that is, although the probability of them occurring simultaneously splits, it does not mean that the events are pairwise independent! Moreover, one cannot define the independence of the events $A, B$ and $C$ by requiring pairwise independence without running into problemsunfortunately the generalisation of Definition 27.1.8 is not that simple, but here we come to the rescue.

Definition 27.1.11 (Mutually independent events). Let $\left(A_{1}, \ldots, A_{n}\right)$ be a collection of events. We say that they are mutually independent if for any subcollection $\left(A_{j_{1}}, \ldots, A_{j_{k}}\right)$ one has

$$
\mathbb{P}\left(\bigcap_{i=1}^{k} A_{j_{i}}\right)=\prod_{i=1}^{k} \mathbb{P}\left(A_{j_{i}}\right) .
$$

### 27.2 Random Variables

Definition 27.2.1 (Random variable). A random variable is a map $X: \Omega \rightarrow \mathcal{T}$, where $\Omega$ is our ambient sample space and $\mathcal{T}$ is the target space-for instance, $\mathbf{R}^{n}$. This random
variable induces a new sample space $\mathcal{X}=X(\Omega)$, on which we can define a new probability function.

Suppose that $\Omega$ is countable. If $\Sigma$ and $\Sigma_{X}$ are $\sigma$-algebras associated to $\Omega$ and $\mathcal{X}$ respectively, then we define $\mathbb{P}_{X}: \Sigma_{X} \rightarrow \mathbf{R}$ given by

$$
\mathbb{P}_{X}(X=x):=\mathbb{P}(\{y \in \Omega: X(y)=x\}),
$$

where $\mathbb{P}: \Sigma \rightarrow \mathbf{R}$ is a probability function. To ease the notation, we shall merely write $\mathbb{P}(X=x)$ rather than $\mathbb{P}_{X}(X=x)$.

Now, if $\Omega$ is an uncountable sample space, we define $\mathbb{P}_{X}$ as follows: for any $A \subseteq X$ let

$$
\mathbb{P}_{X}(X \in A):=\mathbb{P}(\{s \in \Omega: X(s) \in A\}) .
$$

Definition 27.2.2 (Categorical variables). Variables that take a finite set of unordered values are called categorical variables, and are used widely in machine learning contexts.

### 27.3 Distribution Functions

Definition 27.3.1 (Cumulative distribution function). Let $X: \Omega \rightarrow \mathbf{R}$ be a random variable. We define a cumulative distribution function (cdf) associated to $X$ to be map $F_{X}: \Omega \rightarrow \mathbf{R}$ given by

$$
F_{X}(x):=\mathbb{P}_{X}(X \leqslant x)
$$

for each $x \in \Omega$. From its construction, $F_{X}$ is a right-continuous mapping. We shall use the expression $X \sim F_{X}$ to encode " $X$ has a distribution given by $F_{X}$ ".

Theorem 27.3.2 (Cdf necessary and sufficient properties). A map $F: \Omega \rightarrow \mathbf{R}$ is a cumulative distribution if and only if all of the following three conditions hold:
(a) We have limits $\lim _{x \rightarrow-\infty} F(x)=0$ and $\lim _{x \rightarrow \infty} F(x)=1$.
(b) The map $F$ is non-decreasing.
(c) The map $F$ is right-continuous, that is, for any $x_{0} \in \Omega$ we have $\lim _{x \rightarrow+x_{0}} F(x)=$ $F\left(x_{0}\right)$.

Example 27.3.3. Here we list some of the most important cumulative distribution functions:
(a) Let $p$ be the probability of success and $X$ be the number of trials required to obtain a success. The probability functions associated with $X$ is $\mathbb{P}(X=x)=(1-p)^{x-1} p$ and its corresponding cdf is called a geometric distribution:

$$
F_{X}(x)=1-(1-p)^{x} .
$$

Definition 27.3.4. A random variable $X$ is said to be continuous if its associated cumulative distribution function $F_{X}$ is continuous. On the other hand, we say that $X$ is discrete if $F_{X}$ is a step-function.

Definition 27.3.5. Let $\Sigma^{1}$ be the smallest $\sigma$-algebra containing all intervals of real numbers. Two random variables $X$ and $Y$ are said to be identically distributed if for some $A \in \Sigma^{1}$, one has $\mathbb{P}(X \in A)=\mathbb{P}(Y \in A)$. We denote that $X$ and $Y$ have identical distributions by $X \sim Y$.

Theorem 27.3.6. Two random variables $X$ and $Y$ are identically distributed if and only if $F_{X}=F_{Y}$.

## Density \& Mass Functions

Definition 27.3.7 (Probability mass function). The probability mass function (or simply probability function) of a discrete random variable $X$ is a map $f_{X}: \Omega \rightarrow \mathbf{R}$ given by

$$
f_{X}(x)=\mathbb{P}(X=x)
$$

Definition 27.3.8 (Binomial distribution). A random variable $X: \Omega \rightarrow \mathbf{N}$ is said to have a binomial probability function $f$ with parameters $n \in \mathbf{N}$ and $p \in[0,1]$ if

$$
f(x)= \begin{cases}\binom{n}{x} p^{x}(1-p)^{n-x}, & \text { if } 0 \leqslant x \leqslant n \\ 0, & \text { otherwise }\end{cases}
$$

Definition 27.3.9. A random variable $X$ is said to have Bernoulli distribution with parameter $p$ if $X$ is binary-for instance, with values in $\{0,1\}$-and such that $\mathbb{P}(X=$ 1) $=p$.

Definition 27.3.10 (Probability density function). The probability density function (pdf) of a continuous random variable $X$ is a map $f_{X}: \Omega \rightarrow \mathbf{R}$ for which

$$
F_{X}(x)=\int_{-\infty}^{x} f_{X}(t) \mathrm{d} t
$$

Consequently, for any $a<b$ we have

$$
\mathbb{P}(a \leqslant X \leqslant b)=\int_{a}^{b} f_{X}(t) \mathrm{d} t
$$

Remark 27.3.11. In the case of a continuous random variable $X$ we have $\mathbb{P}(X=x)=0$ for any $x$ of the sample space, therefore one has for any interval:

$$
\mathbb{P}(a<X<b)=\mathbb{P}(a<X \leqslant b)=\mathbb{P}(a \leqslant X<b)=\mathbb{P}(a \leqslant X \leqslant b)
$$

Theorem 27.3.12. A map $f_{X}$ is a pdf (or pmf) of a random variable $X$ if and only if the following requirements are met:
(a) The map $f_{X}$ is non-negative.
(b) In the case of a pdf, $\int_{-\infty}^{\infty} f_{X}(x) \mathrm{d} x=1$. On the other hand, for a pmf, $\sum_{x} f_{X}(x)=1$.

Notation 27.3.13. Distributions do not need to depend on a single random variable, when they do we call them univariate, otherwise multivariate distributions. For instance, a multivariate real-valued random variable $X$ is a map $X: \Omega \rightarrow \mathbf{R}^{n}$ having an associated cumulative distribution function and—if existent-probability density functions:

$$
F_{X}(x)=\mathbb{P}\left(X_{1} \leqslant x_{1}, \ldots, X_{n} \leqslant x_{n}\right)=\int_{-\infty}^{x_{1}} \cdots \int_{-\infty}^{x_{n}} f_{X}\left(x_{1}, \ldots, x_{n}\right) \mathrm{d} x_{1} \ldots \mathrm{~d} x_{n}
$$

Definition 27.3.14 (Joint probability). Let $\left(X_{j}: \Omega_{j} \rightarrow \mathbf{R}\right)_{j=1}^{m}$ be a collection of discrete random variables. We define the joint probability of this collection by the probability mass function $f: \prod_{j=1}^{m} \Omega_{j} \rightarrow \mathbf{R}$ given by

$$
f\left(x_{1}, \ldots, x_{m}\right)=\mathbb{P}\left(X_{1}=x_{1}, \ldots, X_{m}=x_{m}\right)=\frac{n\left(x_{1}, \ldots, x_{m}\right)}{N}
$$

where $n\left(x_{1}, \ldots, x_{m}\right)$ is the number of events with state $x_{1}, \ldots, x_{m}$, and $N$ is the total number of events. The probability of $X_{j}=x_{j}$ irrespective of the other random variables as the marginal probability of $X_{j}$ and sometimes denote it by $p\left(x_{j}\right)$. Fixing that $X_{i}=x_{i}$ for each $i \neq j$ we can calculate the conditional probability of $X_{j}=x_{j}$, which we write as $f\left(x_{j} \mid x_{1}, \ldots, x_{m}\right)$.

## Quantile Function

Definition 27.3.15. Let $X: \Omega \rightarrow \mathcal{X}$ be a random variable with a cumulative distribution function $F$. We define the quantile function of $X^{1}$ to be a map Quant: $(0,1) \rightarrow \mathcal{X}$ given by

$$
\operatorname{Quant}(p):=\underset{x \in \mathcal{X}}{\arg \min }(F(x) \geqslant p) .
$$

We call Quant $(p)$ the $p$ quantile of $X$ or the $100 p$ percentile of $X$.

### 27.4 Sum \& Product Rules

Lemma 27.4.1 (Sum rule). Let $X: \Omega_{X} \rightarrow \mathcal{X}$ and $Y: \Omega_{Y} \rightarrow Y$ be multivariate random variables and consider the joint probability $\mathbb{P}: \Omega_{X} \times \Omega_{Y} \rightarrow \mathbf{R}$. The sum rule states that the marginal probability of $X=x$ is given by

$$
\mathbb{P}(x)= \begin{cases}\sum_{y \in y} \mathbb{P}(x, y), & \text { if } y \text { is discrete } \\ \int_{y} \mathbb{P}(x, y) \mathrm{d} y, & \text { if } y \text { is continuous }\end{cases}
$$

Lemma 27.4.2 (Product rule). Given a joint probability $\mathbb{P}: \Omega_{X} \times \Omega_{Y} \rightarrow \mathbf{R}$, the product rule states that

$$
\mathbb{P}(x, y)=\mathbb{P}(y \mid x) \mathbb{P}(x)
$$

[^36]From the product rule we can obtain Bayes' theorem by noting the symmetry $\mathbb{P}(x, y)=\mathbb{P}(x \mid y) \mathbb{P}(y)$, resulting in

$$
\mathbb{P}(x \mid y)=\frac{\mathbb{P}(y \mid x) \mathbb{P}(x)}{\mathbb{P}(y)}
$$

In the eyes of machine learning and Bayesian statistics, this formula gives a way to make inferences about the unobserved random variable $X$ by having a prior knowledge $\mathbb{P}(x)$ and a second random variable $Y$ of which we can observe and have a marginal evidence $\mathbb{P}(y)$ and a relational likelihood $\mathbb{P}(y \mid x)$. This data gives a posterior knowledge $\mathbb{P}(x \mid y)$.

Remark 27.4.3. The value $\mathbb{P}(y \mid x)$ can be called either the "likelihood of $x$ given $y$ " or the "probability of $y$ given $x$ ".

Consider a collection of data $\mathcal{D}$ and model parameters $w$. Frequentists commonly use the maximum likelihood estimator, which aims to find parameters $w$ for which the likelihood function $\mathbb{P}(\mathcal{D} \mid w)$ is maximised-that is, we maximise the probability that, having parameters $w$, the observed set of events is $\mathcal{D}$. In the context of machine learning we normally work with the errors function $\operatorname{err}(w)=-\log (\mathbb{P}(\mathcal{D} \mid w)$ ), and the goal of the learning process is to minimise this error. To be able to work in the frequentist settings, one needs multiple data sets to determine the error bars-one of the methods to deal with this is the bootstrap.

Definition 27.4.4 (Data set bootstrap). Given a data set $X=\left\{x_{1}, \ldots, x_{n}\right\}$, we construct a new data set $X_{j}$ by drawing $n$ points of $X$ at random and with replacement. This procedure can be done $\ell$ times to generate a new collection $\left\{X_{1}, \ldots, X_{\ell}\right\}$ each containing $n$ data points. The statistical accuracy of the parameter estimates will then be evaluated by analysing the variability of predictions between the different bootstrap data sets $X_{j}$.

### 27.5 Functions of Random Variables

Proposition 27.5.1 (The probability function for the discrete case). Let $X: \Omega \rightarrow \mathcal{X}$ be a discrete random variable with probability mass function $f$ and define a random variable $Y:=r(X)$-where $r: \mathcal{X} \rightarrow \boldsymbol{\mathcal { Y }}$ is some function of $X$. Then the probability mass function $g: \boldsymbol{y} \rightarrow \mathbf{R}$ of $Y$ is given by

$$
g(y)=\mathbb{P}(r(X)=y)=\sum_{r(x)=y} f(x)
$$

Proposition 27.5.2 (The probability function for the continuous case). Let $X$ be a continuously distributed random variable with probability density function $f$, and consider a random variable $Y=r(X)$. The cumulative distribution function $G$ of $Y$ is given by

$$
G(y)=\int_{\{x \in \mathcal{X}: r(x) \leqslant y\}} f(x) \mathrm{d} x
$$

Moreover, if $Y$ has a continuous distribution, its probability density function $g$ can be found at a point $y \in \mathcal{Y}$ as a solution of

$$
g(y)=\frac{\mathrm{d} G(y)}{\mathrm{d} y}
$$

assuming that $G$ is differentiable at $y$.
Proposition 27.5.3 (Probability integral transformation). Let $X$ be a continuous random variable with cumulative distribution $F$, and consider a random variable $Y=F X$ such variable is called the probability integral transformation of $X$. Then the cumulative distribution of $Y$ is uniform on the interval [0,1].

Proof. Since $\operatorname{im} F \subseteq[0,1]$ then $\mathbb{P}(Y<0)=0=\mathbb{P}(Y>1)$. From the fact that $F$ is continuous, given any $y \in(0,1)$, we can find a maximum argument

$$
x_{\max }:=\arg \max [F(x)=y]
$$

so that $Y \leqslant y$ if and only if $X \leqslant x_{\max }$. Then the cumulative distribution of $Y$ at $y$ is given by

$$
\mathbb{P}(Y \leqslant y)=\mathbb{P}\left(X \leqslant x_{\max }\right)=F\left(x_{\max }\right)=y
$$

Therefore we conclude that $Y$ has a uniform distribution on the open interval $(0,1)$. Since $Y$ has a continuous distribution, then $Y$ is uniformly distributed in [0, 1]. .

Corollary 27.5.4. Let $Y$ be a uniformly distributed random variable on the interval $[0,1]$. If $F$ is a continuous cumulative distribution function with associated quantile function $q_{F}$, then the random variable $X=q_{F} Y$ has $F$ as its cdf.

Corollary 27.5.5. Let $X$ be a random variable with continuous cdf $F$, and let $G$ be any other continuous cdf. Then the random variable $Z=q_{G} F X$ has $G$ as its cdf.

Proposition 27.5.6. Let $X$ be a random variable with probability distribution $f$ such that $\mathbb{P}(x \in(a, b))=1$. Define a random variable $Y=r X$, where $r$ is an injective and differentiable map on the interval $(a, b)$. Define $(\alpha, \beta):=r((a, b))$ and let $s:(\alpha, \beta) \rightarrow$ $(a, b)$ be the inverse function of $\left.r\right|_{(a, b)}$. Then the probability distribution $g$ of $Y$ is given by

$$
g(y)= \begin{cases}\left|\frac{\mathrm{d} s(y)}{\mathrm{d} y}\right| f s(y), & \text { for } y \in(\alpha, \beta) \\ 0, & \text { otherwise }\end{cases}
$$

Proof. We have two possibilities:
(i) If $r$ is increasing on $(a, b)$ then $s$ is also increasing on $(\alpha, \beta)$. With this, if $G$ is the cumulative distribution of $Y$, we obtain

$$
G(y)=\mathbb{P}(Y \leqslant y)=\mathbb{P}(r(X) \leqslant y)=\mathbb{P}(X \leqslant s(y))=F s(y)
$$

where $F$ is the cumulative distribution for $X$. Therefore, for any $y \in(\alpha, \beta)$ such that $s$ is differentiable at $y$ and $F$ is differentiable at $s(y)$ we have

$$
g(y)=\frac{\mathrm{d} G(y)}{\mathrm{d} y}=\frac{\mathrm{d} F s(y)}{\mathrm{d} y}=f s(y) \frac{\mathrm{d} s}{\mathrm{~d} y} .
$$

Since $s$ is increasing in $(\alpha, \beta)$, then $s^{\prime}$ is positive.
(ii) If $r$ is decreasing on $(a, b)$, then so is $s$ on $(\alpha, \beta)$. Then one has

$$
G(y)=\mathbb{P}(r(X) \leqslant y)=\mathbb{P}(X \geqslant s(y))=1-F s(y) .
$$

Therefore differentiating on each differentiable point we obtain

$$
g(y)=-f s(y) \frac{\mathrm{d} s(y)}{\mathrm{d} y}
$$

From the fact that $s$ is decreasing, then $s^{\prime}$ is negative.
From this we obtain the said equality.
Corollary 27.5.7 (Linear function of random variable). Let $X$ be a random variable with probability distribution $f$, and consider $Y=a X+b$ with $a \neq 0$. The probability density function $g$ of $Y$ is then given by

$$
g(y)=\frac{1}{|a|} f\left(\frac{y-b}{a}\right)
$$

Proposition 27.5.8. Let $\left(X_{1}, \ldots, X_{n}\right)$ be a collection of independent and identically distributed random variables with Bernoulli distribution with parameter $p$. Then the random variable $Y=\sum_{j=0}^{n} X_{j}$ has a binomial distribution with parameters $n$ and $p$.

Proof. Notice that $Y$ is a random variable taking values in $\{0,1,2, \ldots, n\}$, therefore if $Y=y$ there are $\binom{n}{y}$ possible combinations of $\left(X_{1}, \ldots, X_{n}\right)$ in order to have $\sum_{j} X_{j}=y$ since $X_{j} \in\{0,1\}$-where we are going to have exactly $y$ variables evaluating to 1 and the remaining $n-y$ evaluating to 0 . Therefore if $g$ denotes the probability function of $Y$ we have

$$
g(y)=\binom{n}{y} p^{y}(1-p)^{n-y}
$$

Proposition 27.5.9. Let $X$ and $Y$ be random variables with joint probability density function $f$ and let $Z=a_{1} X+a_{2} Y+b$ be a random variable with $a_{1} \neq 0$. Then $Z$ is continuously distributed and has a probability density function given by

$$
g(z)=\int_{-\infty}^{\infty} \frac{1}{\left|a_{1}\right|} f\left(\frac{z-b-a_{2} y}{a_{1}}, y\right) \mathrm{d} y
$$

### 27.6 Expected Value

## Expected Value and Its Properties

Definition 27.6.1 (Expected value). The expected value of a function $g: \mathbf{R} \rightarrow \mathbf{R}$ of a univariate continuous random variable $X: \Omega_{X} \rightarrow \mathcal{X}$ with distribution $f_{X}$ is defined as

$$
\mathbb{E}_{X} g:=\int_{X} g(x) f_{X}(x) \mathrm{d} x
$$

Analogously, if $Y: \Omega_{Y} \rightarrow \boldsymbol{y}$ is a univariate discrete random variable with $Y \sim f_{Y}$, the expected value of $g$ is given by

$$
\mathbb{E}_{Y} g=\sum_{y \in \mathcal{Y}} g(y) f_{Y}(y)
$$

As one might "expect", in the case of an $n$-multivariate random variable $Z: \Omega_{Z} \rightarrow \mathcal{Z}$ we have

$$
\mathbb{E}_{Z} g=\left[\begin{array}{c}
\mathbb{E}_{Z_{1}} g \\
\vdots \\
\mathbb{E}_{Z_{n}} g
\end{array}\right]
$$

We can also define conditional expectation with respect to a conditional distribution, which we shall write as

$$
\mathbb{E}\left[g \mid t_{0}\right]= \begin{cases}\sum_{t \in \mathcal{T}} g(t) f_{T}\left(t \mid t_{0}\right), & \text { if } T \text { is discrete } \\ \int_{\mathcal{T}} g(t) f_{T}\left(t \mid t_{0}\right) \mathrm{d} t, & \text { if } T \text { is continuous }\end{cases}
$$

for a random variable $T: \Omega_{T} \rightarrow \mathcal{T}$ with distribution $f_{T}$.
Corollary 27.6.2. The expected value is a linear map. That is, given random variables $X$, we have

$$
\mathbb{E} Y=a \mathbb{E} X+b
$$

for any pair $a, b \in \mathbf{R}$.
Proposition 27.6.3. Let $X$ be a random variable. If there exists $a \in \mathbf{R}$ such that $\mathbb{P}(X \geqslant a)=1$ then $\mathbb{E} X \geqslant a$, on the other hand, if there exists $b \in \mathbf{R}$ for which $\mathbb{P}(X \leqslant b)=1$ then $\mathbb{E} X \leqslant b$.

Proof. Suppose that $X$ is continuously distributed and has a pdf $f$. If we assume that $\mathbb{P}(X \geqslant a)=1$ then $f(x)=0$ for all $x<a$, therefore

$$
\mathbb{E} X=\int_{a}^{\infty} x f(x) \mathrm{d} x \geqslant \int_{a}^{\infty} a f(x) \mathrm{d} x=a \mathbb{P}(X \geqslant a)=a .
$$

The proof for the upper bound is analogous.
Proposition 27.6.4. Let $X$ be a random variable with $E(X)=a$, and that either $\mathbb{P}(X \geqslant$ $a)=1$ or $\mathbb{P}(X \leqslant a)=1$, then $\mathbb{P}(X=a)=1$.

Proof. Let $X$ be continuously distributed with a pdf $f$, then by assuming that $\mathbb{P}(X \geqslant$ a) $=1$ we get

$$
\begin{aligned}
\mathbb{E} X=\int_{a}^{\infty} x f(x) \mathrm{d} x & =a \mathbb{P}(X=a)+\int_{x>a} x f(x) \mathrm{d} x \\
& \geqslant a \mathbb{P}(X=a)+\int_{x>a} a f(x) \mathrm{d} x \\
& =a
\end{aligned}
$$

Notice that if there exists $x>a$ for which $\mathbb{P}(X=x)>0$ then $\mathbb{E} X>a$, which is a contradiction to the assumption that $\mathbb{E} X=a$, therefore $\mathbb{P}(X>a)=0$ and hence $\mathbb{P}(X=a)=1$.

Proposition 27.6.5 (Additivity of the expected value). Let $\left(X_{1}, \ldots, X_{n}\right)$ be a collection of random variables whose expectation $\mathbb{E} X_{j}$ is finite, then

$$
\mathbb{E}\left(X_{1}+\cdots+X_{n}\right)=\mathbb{E} X_{1}+\cdots+\mathbb{E} X_{n}
$$

Proposition 27.6.6 (Mean of a binomial random variable). Let $X: \Omega \rightarrow \mathbf{N}$ be a binomial random variable with parameters $(n, p)$, then the expected value of $X$ is

$$
\mathbb{E} X=n p
$$

Proof. Let $\left(X_{j}: \Omega \rightarrow\{0,1\}\right)_{j=1}^{n}$ be a collection of independent random variables with a distribution $\mathbb{P}\left(X_{j}=1\right)=p$ and $\mathbb{P}\left(X_{j}=0\right)=1-p$-therefore one has $\mathbb{E} X_{j}=1 \cdot p+0$. $(1-p)=p$. From construction we have $X=\sum_{j=1}^{n} X_{j}$, therefore by Proposition 27.6.5 we obtain $\mathbb{E} X=\sum_{j=1}^{n} \mathbb{E} X_{j}=n p$.

Theorem 27.6.7 (Jensen's inequality). Let $g: \mathbf{R}^{n} \rightarrow \mathbf{R}$ be a convex map ${ }^{2}$, and $X$ be a random vector with finite expected value $\mathbb{E} X$. Then we have the following inequality:

$$
\mathbb{E}(g(X)) \geqslant g(\mathbb{E} X)
$$

Proposition 27.6.8. Let $\left(X_{1}, \ldots, X_{n}\right)$ be a collection of independent random variables with finite expectation $\mathbb{E} X_{j}$, then

$$
\mathbb{E}\left(\prod_{j=1}^{n} X_{j}\right)=\prod_{j=1}^{n} \mathbb{E} X_{j}
$$

[^37]Proof. Let $f$ be the joint pdf of the variables $\left(X_{1}, \ldots, X_{n}\right)$ and denote by $f_{j}$ the marginal pdf of $X_{j}$. Since the variables are independent, it follows that $f=\prod_{j=1}^{n} f_{j}$, therefore

$$
\begin{aligned}
\mathbb{E}\left(\prod_{j=1}^{n} X_{j}\right) & =\int_{-\infty}^{\infty} \cdots \int_{-\infty}^{\infty} f\left(x_{1}, \ldots, x_{n}\right) \prod_{j=1}^{n} x_{j} \mathrm{~d} x_{1} \cdots \mathrm{~d} x_{n} \\
& =\int_{-\infty}^{\infty} \cdots \int_{-\infty}^{\infty} \prod_{j=1}^{n} f_{j}\left(x_{j}\right) x_{j} \mathrm{~d} x_{1} \cdots \mathrm{~d} x_{n} \\
& =\prod_{j=1}^{n} \int_{-\infty}^{\infty} f_{j}\left(x_{j}\right) x_{j} \mathrm{~d} x_{j} \\
& =\prod_{j=1}^{n} \mathbb{E}\left(X_{j}\right) .
\end{aligned}
$$

Proposition 27.6.9. Let $X$ be a non-negative random variable with a cumulative distribution function $F$, then

$$
\mathbb{E} X=\int_{0}^{\infty}(1-F(x)) \mathrm{d} x
$$

## Mean, Median \& Mode

## Univariate Case

Definition 27.6.10 (Mean, median \& mode). We define the mean (or population mean) of an $n$-multivariate random variable $X: \Omega \rightarrow \mathcal{X}$ to be

$$
\mu_{X}=\mathbb{E}_{X} \mathrm{id}=\left[\begin{array}{c}
\mathbb{E}_{X_{1}} \mathrm{id} \\
\vdots \\
\mathbb{E}_{X_{n}} \mathrm{id}
\end{array}\right]
$$

where, as in Definition 27.6 .1 we have for each $1 \leqslant j \leqslant n$ :

$$
\mathbb{E}_{X_{j}} \text { id }= \begin{cases}\int_{X_{j}} x_{j} f_{X}\left(x_{j}\right) \mathrm{d} x_{j}, & \text { if } X \text { is continuous } \\ \sum_{x_{j} \in X_{j}} x_{j} f_{X}\left(x_{j}\right), & \text { if } X \text { is discrete }\end{cases}
$$

The median of a univariate discrete random variable is the middle-most value of the image of the variable. In the case of univariate continuous random variables, we define the median is defined to be the value where the cumulative density function is $1 / 2$.

The mode of a discrete random variable is the value having the highest frequency of occurrence. For continuous random variables, we define the mode as the values corresponding to maximal critical points of the probability density function-which may admit more than a single mode.

Definition 27.6.11 (Covariance, variance \& standard deviation: univariate case). The covariance between two univariate real-valued random variables $X$ and $Y$ is given by

$$
\begin{aligned}
\operatorname{Cov}(X, Y) & :=\mathbb{E}_{X, Y}\left[\left(x-\mu_{X}\right)\left(y-\mu_{Y}\right)\right] \\
& =\mathbb{E}_{X, Y}(x y)-\mu_{X} \mu_{Y}
\end{aligned}
$$

The variance of the random variable $X$ is defined to be

$$
\begin{aligned}
\operatorname{Var} X & :=\operatorname{Cov}(X, X) \\
& =\mathbb{E}_{X}\left[\left(x-\mu_{X}\right)^{2}\right] \\
& =\mathbb{E}_{X}\left(x^{2}\right)-\mu_{X}^{2},
\end{aligned}
$$

which measures how much variability there is in $X$ around its mean value. The standard deviation of $X$ is defined as

$$
\sigma(X):=\sqrt{\operatorname{Var}(X)}
$$

We also define the precision, which is given by

$$
\beta:=\frac{1}{\sigma^{2}}
$$

Proposition 27.6.12. Any univariate random variable has a non-negative variance. Moreover, if $X$ is a bounded univariate random variable, then Var $X$ exists and is finite.
Proof. Let $Y$ be any univariate random variable. Since $\left(Y-\mu_{Y}\right)^{2}$ is non-negative, then $\mathbb{P}\left(\left(Y-\mu_{Y}\right)^{2} \geqslant 0\right)=1$ and by Proposition 27.6.3 we find that Var $Y=\mathbb{E}\left(\left(Y-\mu_{Y}\right)^{2}\right) \geqslant 0$. Now if $X$ is a bounded univariate random variable it follows that $\mathbb{E}(X)=\mu_{X}$ exists and is finite, therefore $\left(X-\mu_{X}\right)^{2}$ is a bounded random variable-therefore $\operatorname{Var}(X)=$ $\mathbb{E}\left(\left(X-\mu_{X}\right)^{2}\right)$ exists and is bounded

Proposition 27.6.13. Let $X$ be a univariate random variable. Then $\operatorname{Var} X=0$ if and only if there exists a value $c$ for which $\mathbb{P}(X=c)=1$.
Proof. Suppose there exists $c$ such that $\mathbb{P}(X=c)=1$, then $\mathbb{P}\left((X-c)^{2}=0\right)=1$ and $\mathbb{E} X=c$. From this we obtain $\operatorname{Var} X=\mathbb{E}\left((X-c)^{2}\right)=0$.

On the other hand, if we assume that $\operatorname{Var} X=0$, then $\mathbb{P}\left((X-\mu)^{2} \geqslant 0\right)=1$ and by Proposition 27.6 .4 we find that $\mathbb{P}\left((X-\mu)^{2}=0\right)=1$, therefore $\mathbb{P}(X=\mu)=1$.

Proposition 27.6.14 (Variance of a linear combination). Let $X$ be a univariate random variable and $Y=a X+b$ for constants $a, b \in \mathbf{R}$, then

$$
\operatorname{Var} Y=a^{2} \operatorname{Var} X \quad \text { and } \quad \sigma_{Y}=|a| \sigma_{X}
$$

Proof. From the linearity of the expected value we have $\mu_{Y}=\mathbb{E}(a X+b)=a \mu_{X}+b$, therefore

$$
\begin{aligned}
\operatorname{Var} Y & =\mathbb{E}\left(\left(Y-\mu_{Y}\right)^{2}\right) \\
& =\mathbb{E}\left(\left(a X+b-a \mu_{X}+b\right)^{2}\right) \\
& =\mathbb{E}\left(a^{2}(X-\mu)^{2}\right) \\
& =a^{2} \operatorname{Var}(X)
\end{aligned}
$$

Proposition 27.6.15 (Additivity of the variance). Let $\left(X_{1}, \ldots, X_{n}\right)$ be a collection of independent random variables having a finite mean, then

$$
\operatorname{Var} \sum_{j=1}^{n} X_{j}=\sum_{j=1}^{n} \operatorname{Var} X_{j}
$$

Therefore, if $\left(a_{1}, \ldots, a_{n}\right) \in \mathbf{R}^{n}$ then

$$
\operatorname{Var} \sum_{j=1}^{n} a_{j} X_{j}=\sum_{j=1}^{n} a_{j}^{2} \operatorname{Var} X_{j}
$$

Proposition 27.6.16 (Variance of a binomial random variable). Let $X: \Omega \rightarrow \mathbf{N}$ be a univariate binomial random variable with parameters $(n, p)$, then

$$
\operatorname{Var} X=n p(1-p)
$$

Proof. Let $\left(X_{j}: \Omega \rightarrow\{0,1\}\right)_{j=1}^{n}$ be a collection of independent random variables with the distribution $\mathbb{P}\left(X_{j}=1\right)=p$ and $\mathbb{P}\left(X_{j}=0\right)=1-p$. It follows from construction that $X=\sum_{j=1}^{n} X_{j}$. Notice now that $X_{j}=X_{j}^{2}$, therefore

$$
\operatorname{Var} X_{j}=\mathbb{E} X_{j}^{2}-\left(\mathbb{E} X_{j}\right)^{2}=p-p^{2}=p(1-p)
$$

From the fact that the variables are independent, using Proposition 27.6.15 we obtain

$$
\operatorname{Var} X=\sum_{j=1}^{n} \operatorname{Var} X_{j}=n p(1-p)
$$

Definition 27.6.17 (Interquartile range). Let $X$ be a univariate random variable with quantile function Quant
Definition 27.6.18 (Correlation). The correlation between two random variables $X$ and $Y$ is given by

$$
\operatorname{Corr}(X, Y):=\frac{\operatorname{Cov}(X, Y)}{\sigma(X) \sigma(Y)} \in[-1,1]
$$

A positive correlation means that if $x$ increases, then $y$ also increases. A negative correlation means that if $x$ increases, then $y$ decreases.

Definition 27.6.19 (Empirical mean \& covariance). Let $\left(x_{j}\right)_{j=1}^{n}$ be a collection of observations (empirical data). We define the empirical mean as

$$
\bar{x}:=\frac{1}{n} \sum_{j=1}^{n} x_{j} .
$$

Analogously, the empirical covariance is defined as

$$
\operatorname{Cov}\left(x_{j}\right)_{j=1}^{n}=\frac{1}{n} \sum_{j=1}^{n}\left\langle x_{j}-\bar{x}, x_{j}-\bar{x}\right\rangle
$$

## Multivariate Case

Definition 27.6.20 (Covariance \& variance: multivariate case). Let $X$ and $Y$ be multivariate random variables with $\operatorname{im} X \subseteq \mathbf{R}^{n}$ and $\operatorname{im} Y \subseteq \mathbf{R}^{m}$. We define the covariance (or cross-variance) between $X$ and $Y$ as

$$
\operatorname{Cov}(X, Y):=\mathbb{E}_{X, Y}(\langle x, y\rangle)-\left\langle\mu_{X}, \mu_{Y}\right\rangle=\operatorname{Cov}(Y, X)^{\top} \in \mathbf{R}^{n} \times \mathbf{R}^{m}
$$

where $\langle-,-\rangle$ is the standard euclidean inner product.
As before, the variance of the multivariate random variable $X$ is given by

$$
\begin{aligned}
\operatorname{Var} X & =\operatorname{Cov}(X, X) \\
& =\mathbb{E}\left(\left\langle x-\mu_{X}, x-\mu_{X}\right\rangle\right) \\
& =\mathbb{E}(\langle x, x\rangle)-\left\langle\mu_{X}, \mu_{X}\right\rangle \\
& =\left[\begin{array}{cccc}
\operatorname{Cov}\left(X_{1}, X_{1}\right) & \operatorname{Cov}\left(X_{1}, X_{2}\right) & \ldots & \operatorname{Cov}\left(X_{1}, X_{n}\right) \\
\vdots & \vdots & \ldots & \vdots \\
\operatorname{Cov}\left(X_{n}, X_{1}\right) & \operatorname{Cov}\left(X_{n}, X_{2}\right) & \ldots & \operatorname{Cov}\left(X_{n}, X_{n}\right)
\end{array}\right]
\end{aligned}
$$

This matrix is symmetric and positive semi-definite.

## Random Variables \& Linear Maps

Lemma 27.6.21. Given a pair of random variables $X$ and $Y$ with states in $\mathbf{R}^{n}$, we can compute the following means:

$$
\begin{aligned}
& \mu_{X+Y}=\mu_{X}+\mu_{Y} \\
& \mu_{X-Y}=\mu_{X}-\mu_{Y}
\end{aligned}
$$

and variances:

$$
\begin{aligned}
& \operatorname{Var}(X+Y)=\operatorname{Var} X+\operatorname{Var} Y+\operatorname{Cov}(X, Y)+\operatorname{Cov}(Y, X) \\
& \operatorname{Var}(X-Y)=\operatorname{Var} X+\operatorname{Var} Y-(\operatorname{Cov}(X, Y)+\operatorname{Cov}(Y, X))
\end{aligned}
$$

Let $Y=A X+b$ be a linear transformation and $X$ a random variable, then $Y$ itself is a random variable. The following relations hold true for the mean of $Y$ :

$$
\mu_{Y}=\mathbb{E}_{Y} Y=\mathbb{E}_{X}[A X+b]=A \mathbb{E}_{X} X+b=A \mu_{X}+b
$$

Moreover, we also have the following relation with the variance of $Y$ :

$$
\begin{aligned}
\operatorname{Var}(Y) & =\operatorname{Var}(A X+b) \\
& =\operatorname{Cov}(A X+b, A X+b) \\
& =\mathbb{E}\left[(A X+b)(A X+b)^{\top}\right]-\mu_{A X+b} \mu_{A X+b}^{\top} \\
& =\mathbb{E}\left[A X X^{\top} A^{\top}+A X b^{\top}+b X^{\top} A^{\top}+b b^{\top}\right]-\left(A \mu_{X} \mu_{X}^{\top} A^{\top}+A \mu_{X} b^{\top}+b \mu_{X}^{\top} A^{\top}-b b^{\top}\right) \\
& =A \mathbb{E}\left[X X^{\top}\right] A^{\top}-A \mu_{X} \mu_{X}^{\top} A^{\top} \\
& =\operatorname{Var}(A X) \\
& =A \operatorname{Var}(X) A^{\top}
\end{aligned}
$$

Finally, we can compute the covariance of $X$ and $Y$ as

$$
\begin{aligned}
\operatorname{Cov}(X, Y) & =\mathbb{E}\left[X(A x+b)^{\top}\right]-\mu_{X} \mu_{A X+b}^{\top} \\
& =\mu_{X} b^{\top}+\mathbb{E}\left[X X^{\top}\right] A^{\top}-\mu_{X} b^{\top}-\mu_{X} \mu_{X}^{\top} A^{\top} \\
& =\mathbb{E}\left[X X^{\top}\right] A^{\top}-\mu_{X} \mu_{X}^{\top} A^{\top} \\
& =\operatorname{Var}(X) A^{\top}
\end{aligned}
$$

Lemma 27.6.22. If $X$ and $Y$ are statistically independent random variables, then
(a) $\operatorname{Var}(X+Y)=\operatorname{Var} X+\operatorname{Var} Y$.
(b) $\operatorname{Cov}(X, Y)=0$.

It is to be noted that two random variables may have a zero covariance but still be statistically dependent-this is because covariance merely measures linear dependence.

Definition 27.6.23 (Conditional independence). Two random variables $X$ and $Y$ are said to be conditionally independent of a given random variable $Z$ when

$$
\mathbb{P}(x, y \mid z)=\mathbb{P}(x \mid z) \mathbb{P}(y \mid z)
$$

for all states $z$ of $Z$. If we use Lemma 27.4.2 we obtain $\mathbb{P}(x, y \mid z)=\mathbb{P}(x \mid y, z) \mathbb{P}(y \mid z)$ and hence the conditional independence can be reformulated as

$$
\mathbb{P}(x \mid y, z)=\mathbb{P}(x \mid z)
$$

Definition 27.6.24 (Inner product of random variables). Given two random variables $X$ and $Y$, we can define their inner product to be

$$
\langle X, Y\rangle:=\mathbb{E}(X Y)
$$

Also if it is the case that $\mu_{X}=0=\mu_{Y}$, then in particular $\langle X, Y\rangle=\operatorname{Cov}(X, Y)$ and we would have

$$
\operatorname{Corr}(X, Y)=\frac{\operatorname{Cov}(X, Y)}{\sqrt{\operatorname{Var}(X) \operatorname{Var}(Y)}}=\frac{\langle X, Y\rangle}{\|X\|\|Y\|}=\cos \theta
$$

where $\theta$ is the angle between the vectors $X$ and $Y$.

## Moments

Definition 27.6.25 (Moment). Given a random variable $X$ and a positive integer $n$, the $n$-th moment of $X$ is defined to be $\mathbb{E}\left(X^{n}\right)$.

Corollary 27.6.26. The $n$-th moment of a random variable $X$ exists if and only if $\mathbb{E}|X|^{n}<\infty$. Moreover, if $X$ is bounded, then all moments of $X$ do exist (which is a sufficient, but not necessary condition).

Corollary 27.6.27. If the $k$-th moment of a random variable $X$ exists, then for any $0<j<k$ the $j$-th moment of $X$ also exists.

Proof. Let $f$ be a probability density function over $X$, then

$$
\begin{aligned}
\mathbb{E}|X|^{j} & =\int_{-\infty}^{\infty}|x|^{j} f(x) \mathrm{d} x \\
& =\int_{-\infty}^{1}|x|^{j} f(x) \mathrm{d} x+\int_{1}^{\infty}|x|^{j} f(x) \mathrm{d} x \\
& \leqslant \int_{-\infty}^{1} f(x) \mathrm{d} x+\int_{1}^{\infty}|x|^{k} f(x) \mathrm{d} x \\
& =\mathbb{P}(|X| \leqslant 1)+\mathbb{E}|X|^{k} .
\end{aligned}
$$

Now since the $k$-th moment is assumed to exist, then $\mathbb{E}|X|^{k}<\infty$ and since $\mathbb{P}(|X| \leqslant$ 1) $\in[0,1]$, then $\mathbb{E}|X|^{j}<\infty$.

Corollary 27.6.28. If the second moment of $X$ exists, then $X$ has both a mean and a variance.

Definition 27.6.29 (Central moment). Given a random variable $X$ with mean $\mathbb{E} X=\mu$, the $n$-th central moment of $X$ is defined to be the expectation $\mathbb{E}\left((X-\mu)^{n}\right)$.

Corollary 27.6.30. If $X$ is symmetric with respect to its mean $\mu$, then for any odd positive integer $k$ such that the $k$-th central moment of $X$ exists, then $\mathbb{E}\left((X-\mu)^{k}\right)=0$.

Definition 27.6.31 (Random variable skewness). Let $X$ be a random variable having mean $\mu$ and standard deviation $\sigma$. If $X$ has a finite third moment, we define the skewness of $X$ to be the value

$$
\frac{\mathbb{E}\left((X-\mu)^{3}\right)}{\sigma}
$$

measuring the lack of symmetry of $X$.

## Moment Generating Functions

Definition 27.6.32 (Moment generating function). Let $X$ be a random variable. We define a moment generating function (mgf) of $X$ to be the map $\psi: \mathbf{R} \rightarrow \mathbf{R}$ given by

$$
\psi(t):=\mathbb{E} e^{t X}
$$

Since $X$ may not be bounded, $\psi$ may not be continuous since the expectation can be infinite in some points. The mgf depends solely on the distribution underlying $X$, hence if $X$ and $Y$ has the same distribution of $X$, their moment generating functions will coincide.

Proposition 27.6.33. Let $X$ be a random variable whose moment generating function $\psi$ is finite in an open interval $(-\varepsilon, \varepsilon)$ around zero. Then for any positive integer $n$ the $n$-th moment of $X$ is finite and equals the $n$-th derivative of $\psi$ at zero:

$$
\mathbb{E} X^{n}=\psi^{(n)}(0)
$$

In particular, we have $\mu=\psi^{\prime}(0)$ and $\operatorname{Var}(X)=\psi^{\prime \prime}(0)-\psi^{\prime}(0)^{2}$.

Proposition 27.6.34. Given a random variable $X$ with an associated $\operatorname{mgf} \psi_{X}$. If $Y=$ $a X+b$ is another random variable, with an mgf $\psi_{Y}$, then

$$
\psi_{Y}(t)=e^{b t}+\psi_{X}(a t)
$$

Proposition 27.6.35. Let $\left(X_{j}, \psi_{j}\right)_{j=1}^{n}$ be a collection of independent random variables together with their associated mgf. Then if $Y=\sum_{j=1}^{n} X_{j}$ it follows that the mgf $\psi$ of $Y$ is given by

$$
\psi=\prod_{j=1}^{n} \psi_{j} .
$$

Proof. Notice that $\psi(t)=\mathbb{E} e^{t \sum_{j=1}^{n} X_{j}}=\mathbb{E}\left(\prod_{j=1}^{n} e^{t X_{j}}\right)$ and since the random variables are independent, the expected value preserves the product: $\psi(t)=\prod_{j=1}^{n} \mathbb{E} e^{t X_{j}}$. 白
Theorem 27.6.36. If $X$ and $Y$ are two random variables whose mgf's agree in an open interval $(-\varepsilon, \varepsilon)$ around zero, then the probability distributions of $X$ and $Y$ are identical.

Proof.
Proof.

## Binomial Distribution \& MGF's

Proposition 27.6.37 (Binomial distribution mgf). If $X: \Omega \rightarrow \mathbf{N}$ is a binomial random variable with parameters $(n, p)$, then the mgf associated with $X$ is

$$
\psi(t)=\left(p e^{t}+1-p\right)^{n}
$$

Proof. Let $\left(X_{j}: \Omega \rightarrow\{0,1\}\right)_{j=1}^{n}$ be a collection of independent random variables whose distribution is $\mathbb{P}\left(X_{j}=1\right)=p$ and $\mathbb{P}\left(X_{j}=0\right)=1-p$, so that $X=\sum_{j=1}^{n} X_{j}$. Since each $X_{j}$ has the same distribution, their moment generating function is the same $\psi^{\prime}$ :

$$
\psi^{\prime}(t)=\mathbb{E}\left(e^{t X_{j}}\right)=e^{t} \cdot \mathbb{P}\left(X_{j}=1\right)+1 \cdot \mathbb{P}\left(X_{j}=0\right)=e^{t} p+1-p
$$

Therefore from Proposition 27.6 .35 we obtain

$$
\psi(t)=\psi^{\prime}(t)^{n}=\left(e^{t} p+1-p\right)^{n}
$$

Theorem 27.6.38 (Additivity of binomial random variables). If $X$ and $Y$ are independent binomial random variables with parameters $(n, p)$ and $(m, p)$ respectively, then the random variable $X+Y$ is binomial and has parameters $(n+m, p)$.

Proof. Let $\psi_{X}$ and $\psi_{Y}$ be the mgf's of $X$ and $Y$ respectively. From Proposition 27.6.37 and Proposition 27.6.35 we know that the mgf $\psi$ of $X+Y$ is given by

$$
\psi(t)=\psi_{X}(t) \psi_{Y}(t)=\left(p e^{t}+1-p\right)^{n+m}
$$

which is also the mgf of the binomial random variable whose parameters are $(n+$ $m, p)$ —therefore by Theorem 27.6.36 we can conclude that $X+Y$ is in fact the binomial random variable with parameters $(n+m, p)$.

### 27.7 Distributions

Definition 27.7.1 (Gaussian distribution). Let $X$ be a univariate random variable with mean $\mu$ and standard deviation $\sigma$. The Gaussian distribution of $X$ has a density function given by

$$
\mathcal{N}\left(x \mid \mu, \sigma^{2}\right):=\frac{1}{\sqrt{2 \pi \sigma^{2}}} \exp \left(-\frac{(x-\mu)^{2}}{2 \sigma^{2}}\right)
$$

Now, if $X$ where an $n$-multivariate random variable, then the Gaussian distribution of $X$ would have a density function defined as

$$
\mathcal{N}(x \mid \mu, \operatorname{Var} X):=\frac{1}{\sqrt{(2 \pi)^{n} \operatorname{det}(\operatorname{Var} X)}} \exp \left(-\frac{1}{2}(x-\mu)^{\top} \operatorname{Var}(X)^{-1}(x-\mu)\right)
$$

The special case where $\mu=0$ and $\operatorname{Var}(X)=$ id is referred to as the standard normal distribution.

Proposition 27.7.2 (Gaussian distribution properties). The following items are properties pertaining to the Gaussian distribution $\mathcal{N}$ :
(a) The Gaussian distribution is normalised.
(b) The maximum of a Gaussian distribution, its mode, coincides with the mean.
(c) The likelihood function associated with a finite dataset of scalar observations $\mathcal{D}$ is given by

$$
\mathbb{P}\left(\mathcal{D} \mid \mu, \sigma^{2}\right)=\prod_{x \in \mathcal{D}} \mathcal{N}\left(x \mid \mu, \sigma^{2}\right)
$$

Remark 27.7.3 (Finding $\mu$ and $\sigma^{2}$ with the maximum likelihood). In order to determine the unknown parameters $\mu$ and $\sigma^{2}$ associated with the dataset $\mathcal{D}$ we can try to find the parameters that maximise the likelihood function-or, equivalently, maximising the function

$$
\begin{equation*}
\log \mathbb{P}\left(\mathcal{D} \mid \mu, \sigma^{2}\right)=-\frac{1}{2 \sigma^{2}} \sum_{x \in \mathcal{D}}(x-\mu)^{2}-\frac{N}{2}\left(\log \left(\sigma^{2}\right)+\log (2 \pi)\right) \tag{27.1}
\end{equation*}
$$

where $N:=|\mathcal{D}|$ is the total number of data points. Maximising Eq. (27.1) with respect to $\mu$ gives

$$
\mu_{\mathrm{ML}}=\frac{1}{N} \sum_{x \in \mathcal{D}} x
$$

known as the sample mean, while maximising with respect to $\sigma^{2}$ we obtain

$$
\sigma_{\mathrm{ML}}^{2}=\frac{1}{N} \sum_{x \in \mathcal{D}}\left(x-\mu_{\mathrm{ML}}\right)^{2}
$$

called the sample variance with respect to the sample mean $\mu_{\mathrm{ML}}$.

The problem with this approach, however, is that the obtained sample variance is underestimated with respect to the true variance $\sigma^{2}$. Indeed, one has

$$
\mathbb{E}\left[\sigma_{\mathrm{ML}}^{2}\right]=\frac{N-1}{N} \sigma^{2}
$$

while we obtain a correct mean $\mathbb{E}\left[\mu_{\mathrm{ML}}\right]=\mu$. This effect is called bias. In order to deal with that, we can calculate an unbiased variance by

$$
\widetilde{\sigma}^{2}=\frac{N}{N-1} \sigma_{\mathrm{ML}}^{2}=\frac{1}{N-1} \sum_{x \in \mathcal{D}}\left(x-\mu_{\mathrm{ML}}\right)^{2}
$$

Definition 27.7.4 (Hyperparameters). Parameters that control the distribution of model parameters are called hyperparameters.

Let $X$ and $Y$ be multivariate random variables and consider the Gaussian distribution

$$
f(x, y)=\mathcal{N}\left(\left[\begin{array}{l}
\mu_{X} \\
\mu_{Y}
\end{array}\right],\left[\begin{array}{cc}
\operatorname{Var} X & \operatorname{Cov}(X, Y) \\
\operatorname{Cov}(Y, X) & \operatorname{Var} Y
\end{array}\right]\right)
$$

### 27.8 Model Selection

Definition 27.8.1 ( $S$-fold cross-validation). Let $N$ be the number of data points available in a given dataset. The $S$-fold cross-validation for evaluating model performance consist in creating $S \leqslant N^{3}$ distinct groups of the available data-then $S-1$ of those are used as training data for the model, while the remaining group is used as a validation set. We permute the validation set across all possible $S$ groups and finally calculate the model performance as the average score from the $S$ total runs.

## Loss Functions

In the case of classification problems, we may analyse the number of misclassifications via what is called a loss matrix. Let $C=\left\{C_{k}\right\}_{k=1}^{n}$ be the classes associated with a dataset $\mathcal{D}$ and define the loss matrix $L$ to be an $n \times n$ matrix whose entries $L_{i j}$ are the number of data points of true class $C_{i}$ which where predicted to pertain to class $C_{j}$. For each input $x$ the uncertainty associated with the true class of $x$ is given by the probability distribution $\mathbb{P}\left(x, C_{k}\right)$. Our goal in developing classification algorithms will be to minimise the average loss computed with respect to the associated uncertainty over the decision spaces ${ }^{4}\left\{R_{k}\right\}_{k=1}^{n}$ associated with each class $C_{j}$ :

$$
\mathbb{E}[L]=\sum_{i=1}^{n} \sum_{j=1}^{n} \int_{R_{j}} L_{i j} \mathbb{P}\left(x, C_{i}\right) \mathrm{d} x
$$

[^38]Since $\mathbb{P}\left(x, C_{k}\right)=\mathbb{P}\left(C_{k} \mid x\right) \mathbb{P}(x)$ then the decision rule that minimises the expected loss is the one assigning $x$ to the class $C_{j_{0}}$ where

$$
j_{0}=\underset{j}{\arg \min } \sum_{i=1}^{n} L_{i j} \mathbb{P}\left(C_{i} \mid x\right) .
$$

In the case of regression problems, given a dataset $\mathcal{D}$ one wants to create the best possible estimate function $g: \mathbf{R}^{d} \rightarrow \mathbf{R}^{m}$ for the unknown map $f: \mathbf{R}^{d} \rightarrow \mathbf{R}^{m}$ associated with the true values of $\mathcal{D}$. To that end, we must choose a loss function

$$
L: f(\mathcal{D}) \times g(\mathcal{D}) \rightarrow \mathbf{R}
$$

for instance, the squared loss is given by $L(f(x), g(x))=(g(x)-f(x))^{2}$. The average of the expected loss, which we want to minimise, is given by

$$
\mathbb{E}[L]=\iint L(t, g(x)) \mathbb{P}(x, t) \mathrm{d} x \mathrm{~d} t
$$

## Chapter 28

## Learning

### 28.1 Learnability

Definition 28.1.1 (True error). Consider a target function $f: \mathcal{X} \rightarrow \mathcal{Y}$ and a probability distribution $\mathcal{D}$ over $\mathcal{X}$. If $h: \mathcal{X} \rightarrow \mathcal{Y}$ is a prediction rule for the learning problem $(\mathcal{D}, f)$, we define the true error of $h$ as

$$
L_{(\mathcal{D}, f)}(h):=\mathbb{P}_{x \sim \mathcal{D}}(h(x) \neq f(x))=\mathcal{D}(\{x \in \mathcal{X}: h(x) \neq f(x)\}),
$$

which is the probability of sampling a point $x \in \mathcal{X}$, according to the distribution $\mathcal{D}$, on which the predictor $h$ fails to meet $f$.

Definition 28.1.2 (Training error). Let $S=\left(x_{i}, y_{i}\right)_{i=1}^{m}$ be a training set, sampled from an unknown probability distribution $\mathcal{D}$ and labelled by a target function $f: \mathcal{X} \rightarrow \mathcal{Y}$. Let $h_{S}: \mathcal{X} \rightarrow \mathcal{Y}$ be the hypothesis returned by a learning algorithm based on the training set $S$. We define the training error $L_{S}: \mathcal{Y}^{\mathcal{X}} \rightarrow \mathbf{R}$ to be the map

$$
L_{S}(h):=\frac{\left|\left\{i \in[m]: h\left(x_{i}\right) \neq y_{i}\right\}\right|}{m} .
$$

Definition 28.1.3 (Empirical risk minimisation). The learning paradigm known as empirical risk minimisation (ERM) is defined to have the goal of minimising the training error. As this may cause overfitting, we restrict the set of possible predictors to a hypothesis collection $\mathcal{H}$-this approach induces a bias to the model. Given a training sample $S$, we define the learner $\mathrm{ERM}_{\mathcal{H}}: 2^{X} \rightarrow \boldsymbol{y}^{X}$ to choose a predictor ERM $\mathcal{H}_{\mathcal{H}}(S):=h_{S}$ such that

$$
h_{S} \in \underset{h \in \mathcal{H}}{\arg \min } L_{S}(h)
$$

Definition 28.1.4 (Realisability assumption). Within the hypothesis collection $\mathcal{H}$ there exists $h^{\star}$ for which the training error is null:

$$
L_{S}\left(h^{\star}\right)=0 .
$$

Definition 28.1.5 (Independently identically distributed assumption (i.i.d.)). We shall assume that the training set $S$ has points which are independently and identically distributed (i.i.d.) with respect to the probability distribution $\mathcal{D}$, shortly we write that $S \sim \mathcal{D}^{|S|}$.

Since $S$ is chosen by a randomised procedure, the predictor $h_{S}$ and risk $L_{(\mathcal{D}, f)}\left(h_{S}\right)$ are both random variables. This allows us to talk about the probability of $L_{(\mathcal{D}, f)}\left(h_{S}\right)$ not being large.
Definition 28.1.6 (Confidence parameter). We shall denote by $\delta$ the probability of $S$ being a non-representative sample of $\mathcal{X}$ with respect to the distribution $\mathcal{D}$, and $1-\delta$ the confidence parameter of the predictor $h_{S}$.

Definition 28.1.7 (Accuracy parameter). We denote by $\varepsilon$ the accuracy parameter, which measures how large is $L_{(\mathcal{D}, f)}\left(h_{S}\right)$ as follows: if $L_{(\mathcal{D}, f)}\left(h_{S}\right)>\varepsilon$ we interpret the result of the algorithm as a failure, while $L_{(\mathcal{D}, f)}\left(h_{S}\right) \leqslant \varepsilon$ as a approximately correct predictor.

Proposition 28.1.8 (Failure bound). Consider a learner with a finite hypothesis class $\mathcal{H}$. The probability of obtaining a sample $S$, of size $m$, on which the learner fails the accuracy parameter is bound by

$$
\mathcal{D}^{m}\left(\left\{\pi_{\mathcal{X}} S: L_{(\mathcal{D}, f)}\left(h_{S}\right)>\varepsilon\right\}\right) \leqslant|\mathcal{H}| e^{-\varepsilon m}
$$

where $\pi_{\chi}$ is the projection of the first coordinate.
Proof. Let $\mathcal{H}_{\text {bad }}:=\left\{h \in \mathcal{H}: L_{(\mathcal{D}, f)}>\varepsilon\right\}$ be the collection of all hypothesis which fail to meet the expected accuracy of the learner, and define

$$
M:=\left\{\pi_{\mathcal{X}} S: \exists h \in \mathcal{H}_{\mathrm{bad}}, L_{S}(h)=0\right\}=\bigcup_{h \in \mathcal{H}_{\text {bad }}}\left\{\pi_{\mathcal{X}} S: L_{S}(h)=0\right\}
$$

the collection of all possible misleading non-representative samples for which the training error can be nullified by the learner.

The realisability assumption says that $L_{S}\left(h_{S}\right)=0$, therefore $L_{(\mathcal{D}, f)}\left(h_{S}\right)>\varepsilon$ is only possible in the situation where there exists a predictor $h \in \mathcal{H}_{\text {bad }}$ such that $L_{S}(h)=0$, that is

$$
\left\{\pi \mathcal{X} S: L_{(\mathcal{D}, f)}\left(h_{S}\right)>\varepsilon\right\} \subseteq M
$$

Therefore, by the union bound applied to the distribution $\mathcal{D}^{m}$ we have

$$
\begin{aligned}
\mathcal{D}^{m}\left(\left\{\pi_{\mathcal{X}} S: L_{(\mathcal{D}, f)}\left(h_{S}\right)>\varepsilon\right\}\right) & \leqslant \mathcal{D}^{m}(M) \\
& =\mathcal{D}^{m}\left(\bigcup_{h \in \mathcal{H}_{\text {bad }}}\left\{\pi_{\mathcal{X}} S: L_{S}(h)=0\right\}\right) \\
& \leqslant \sum_{h \in \mathcal{H}_{\text {bad }}} \mathcal{D}^{m}\left(\left\{\pi_{\mathcal{X}} S: L_{S}(h)=0\right\}\right)
\end{aligned}
$$

Now notice that for any $h \in \mathcal{H}_{\text {bad }}$ we have

$$
\left\{\pi \mathcal{X} S: L_{S}(h)=0\right\}=\left\{\pi \mathcal{X} S: \forall j \in[m], h\left(x_{j}\right)=f\left(x_{j}\right)\right\}
$$

where $\pi_{\mathcal{X}} S=\left(x_{j}\right)_{j=1}^{m}$. Since each point $x_{j} \in \mathcal{X}$ composing $S$ is chosen in an independently and identically distributed fashion, it follows that

$$
\begin{aligned}
\mathcal{D}^{m}\left(\left\{\pi \mathcal{X} S: L_{S}(h)=0\right\}\right) & =\mathcal{D}^{m}\left(\left\{\pi \mathcal{X} S: \forall j \in[m], h\left(x_{j}\right)=f\left(x_{j}\right)\right\}\right) \\
& =\prod_{j=1}^{m} \mathcal{D}\left(\left\{x_{j} \in \mathcal{X}: h\left(x_{j}\right)=f\left(x_{j}\right)\right\}\right) .
\end{aligned}
$$

Moreover, since we define $L_{(\mathcal{D}, f)}=\mathcal{D}(\{x: h(x) \neq f(x)\})$ we obtain for each $j \in[m]$ the bound

$$
\mathcal{D}\left(\left\{x_{j}: h\left(x_{j}\right)=f\left(x_{j}\right)\right\}\right)=1-L_{(\mathcal{D}, f)}(h) \leqslant 1-\varepsilon
$$

since $L_{(\mathcal{D}, f)}(h)>\varepsilon$ for any $h \in \mathcal{H}_{\text {bad }}$. Now if we use the inequality $1-\varepsilon \leqslant e^{-\varepsilon}$ we obtain

$$
\mathcal{D}^{m}\left(\left\{\pi_{\mathcal{X}} S: L_{S}(h)=0\right\}\right) \leqslant(1-\varepsilon)^{m} \leqslant e^{-m \varepsilon}
$$

From this analysis we can conclude that

$$
\mathcal{D}^{m}\left(\left\{\pi_{X} S: L_{(\mathcal{D}, f)}(h)>\varepsilon\right\}\right) \leqslant\left|\mathcal{H}_{\mathrm{bad}}\right| e^{-m \varepsilon} \leqslant|\mathcal{H}| e^{-m \varepsilon}
$$

which is the bound that we settled to prove.
Corollary 28.1.9. Let $\mathcal{H}$ be a finite hypothesis collection, and consider parameters $0<\delta<1$ and $\varepsilon>0$. If $m \in \mathbf{N}$ is such that

$$
m \geqslant \frac{\log (|\mathcal{H}| / \delta)}{\varepsilon}
$$

then for any labelling map $f: \mathcal{X} \rightarrow \mathcal{Y}$ together with a distribution $\mathcal{D}$ following the realisability assumption, we have a probability of at least $1-\delta$ of choosing an i.i.d. sample $S$ with size $m$ for which every ERM hypothesis $h_{S}$ satisfies

$$
L_{(\mathcal{D}, f)}\left(h_{S}\right) \leqslant \varepsilon
$$

In other words, for a sufficiently large $m$, the learner $\mathrm{ERM}_{\mathcal{H}}$ will be probably (with a confidence of at least $1-\delta$ ) approximately (up to an error of $\varepsilon$ ) correct (PAC).

### 28.2 PAC Learning Model

Definition 28.2.1 (PAC learning). A hypothesis class $\mathcal{H}$ is said to be PAC learnable if there exists a map

$$
m_{\mathcal{H}}:(0,1)^{2} \longrightarrow \mathbf{N}
$$

and a learning algorithm $A$ such that for every tuple $(\varepsilon, \delta, f, \mathcal{D}, S)$ where

- $\varepsilon, \delta \in(0,1)$ are the accuracy and confidence parameters, respectively.
- $f: \mathcal{X} \rightarrow\{0,1\}$ is a binary labelling map on $\mathcal{X}$.
- $\mathcal{D}$ is a probability distribution on $\mathcal{X}$.
- The realisability assumption holds with respect to the triple $(\mathcal{H}, \mathcal{D}, f)$.
- The set $S=\left(x_{j}, f\left(x_{j}\right)\right)_{j=1}^{m}$ with $m \geqslant m_{\mathcal{H}}(\varepsilon, \delta)$ is composed of i.i.d. samples of $\mathcal{X}$ generated by $\mathcal{D}$ and labelled by $f$.

The algorithm $A$ returns a predictor $A(S)=h$ with a probability of at least $1-\delta$, over the choice of $S$, such that

$$
L_{(\mathcal{D}, f)}(h) \leqslant \varepsilon
$$

The $\operatorname{map} m_{\mathcal{H}}:(0,1)^{2} \rightarrow \mathbf{N}$ determines the sample complexity necessary to ensure that the learning algorithm will result in a PAC solution. We shall impose $m_{\mathcal{H}}$ to return the minimal sample complexity such that $\mathcal{H}$ is PAC learnable with accuracy $\varepsilon$ and confidence $\delta$.

Rephrasing Corollary 28.1 .9 with our new jargon:
Corollary 28.2.2. Every finite hypothesis class $\mathcal{H}$ is PAC learnable with sample complexity

$$
m_{\mathcal{H}}(\varepsilon, \delta) \leqslant\left\lceil\frac{\log (|\mathcal{H}| / \delta)}{\varepsilon}\right\rceil .
$$

## Agnostic PAC Learning Model

Let $\mathcal{X}$ be our domain space of features and $\mathcal{Y}$ be the space of labels. We shall now consider $\mathcal{D}$ to be the joint distribution over the product space $\mathcal{X} \times \boldsymbol{y}$. This allows for two samples corresponding to the same point $x \in \mathcal{X}$ to assume different labels $y_{1}, y_{2} \in \boldsymbol{y}$. With this in mind, we need to revise the true error to be in accordance with this new distribution $\mathcal{D}$.

Definition 28.2.3 (Revising the true error). Given a probability distribution $\mathcal{D}$ over $\mathcal{X} \times \mathcal{Y}$, and a predictor $h: \mathcal{X} \rightarrow \mathcal{Y}$, the true error of $h$ is given by

$$
L_{\mathcal{D}}(h):=\mathbb{P}_{(x, y) \sim \mathcal{D}}(h(x) \neq y)=\mathcal{D}(\{(x, y) \in \mathcal{X} \times \mathcal{Y}: h(x) \neq y\}) .
$$

Proposition 28.2.4 (Bayes optimal predictor). Let $\mathcal{D}$ be any distribution on $\mathcal{X} \times\{0,1\}$. The best predictor $f_{\mathcal{D}}: \mathcal{X} \rightarrow\{0,1\}$, called the Bayes optimal predictor ${ }^{1}$, will be given by

$$
f_{\mathcal{D}}(x):= \begin{cases}1, & \text { if } \mathbb{P}_{(x, y) \sim \mathcal{D}}(y=1 \mid x) \geqslant 1 / 2 \\ 0, & \text { otherwise }\end{cases}
$$

That is, for any predictor $h: \mathcal{X} \rightarrow\{0,1\}$ we have $L_{\mathcal{D}}\left(f_{\mathcal{D}}\right) \leqslant L_{\mathcal{D}}(h)$.
Definition 28.2.5 (Agnostic PAC learning model). A hypothesis class $\mathcal{H}$ is said to be agnostic PAC learnable if there exists a minimal map $m_{\mathcal{H}}:(0,1)^{2} \rightarrow \mathbf{N}$ and leaning algorithm $A$ such that for every tuple $(\varepsilon, \delta, \mathcal{D}, S)$ where

- $\varepsilon, \delta \in(0,1)$ are the accuracy and confidence parameters, respectively.
- $\mathcal{D}$ is a probability distribution over $\mathcal{X} \times \mathcal{Y}$.
- $S \in(\mathcal{X} \times \boldsymbol{Y})^{m}$ is an i.i.d. sample of size $m \geqslant m_{\mathcal{H}}(\varepsilon, \delta)$ generated by $\mathcal{D}$.

The algorithm $A$ returns a predictor $A(S):=g \in \mathcal{H}$ with a probability of at least $1-\delta$ over the choice of $S$, such that

$$
L_{\mathcal{D}}(g) \leqslant \min _{h \in \mathcal{H}} L_{\mathcal{D}}(h)+\varepsilon .
$$

[^39]
### 28.3 Extending The PAC Model

Definition 28.3.1 (Generalised loss function). Let $\mathcal{H}$ be a collection of hypothesis, and $\mathcal{Z}$ a domain of interest. Any map of the type

$$
\ell: \mathcal{H} \times \mathcal{Z} \rightarrow \mathbf{R}_{\geqslant 0}
$$

is a loss function.
Definition 28.3.2 (Generalised error functions). Consider a hypothesis class $\mathcal{H}$, a domain $\mathcal{Z}$ with associated distribution $\mathcal{D}$, and a loss function $\ell: \mathcal{H} \times \mathcal{Z} \rightarrow \mathbf{R}_{\geqslant 0}$. We define the following:
(a) The true error of a predictor $h \in \mathcal{H}$ is given by

$$
L_{\mathcal{D}}(h):=\mathbb{E}_{z \sim \mathcal{D}}[\ell(h, z)] .
$$

(b) Given an i.i.d. sample $S \in \mathcal{Z}^{m}$, we define the empirical error of a predictor $h \in \mathcal{H}$ to be

$$
L_{S}(h):=\frac{1}{m} \sum_{z \in S} \ell(h, z)
$$

Example 28.3.3. The following are extensively used loss functions:
(a) (Classifier loss) Given a domain $\mathcal{Z}=\mathcal{X} \times \mathcal{Y}$ and a predictor $h$, we define the classifier loss function

$$
\ell_{\text {class }}(h,(x, y)):= \begin{cases}0, & \text { if } h(x)=y \\ 1, & \text { if } h(x) \neq y\end{cases}
$$

(b) (Square loss) Consider a domain $\mathcal{Z}=\mathcal{X} \times \mathcal{Y}$ and a predictor $h$, we define the square loss function as

$$
\ell_{\mathrm{sq}}(h,(x, y)):=(h(x)-y)^{2},
$$

which is commonly used in regression learning problems.
Definition 28.3.4 (Agnostic PAC learning model for generalised loss functions). A hypothesis class $\mathcal{H}$ is said to be agnostic PAC learnable with respect to a domain $\mathcal{Z}$ and loss function $\ell: \mathcal{H} \times \mathcal{Z} \rightarrow \mathbf{R}_{\geqslant 0}$ if there exists a minimal map $m_{\mathcal{H}}:(0,1)^{2} \rightarrow \mathbf{N}$ and leaning algorithm $A$ such that for every tuple $(\varepsilon, \delta, \mathcal{D}, S)$ where

- $\varepsilon, \delta \in(0,1)$ are the accuracy and confidence parameters, respectively.
- $\mathcal{D}$ is a probability distribution over $\mathcal{Z}$.
- $S \in \mathcal{Z}^{m}$ is an i.i.d. sample of size $m \geqslant m_{\mathcal{H}}(\varepsilon, \delta)$ generated by $\mathcal{D}$.

The algorithm $A$ returns a predictor $A(S):=g \in \mathcal{H}$ with a probability of at least $1-\delta$ over the choice of $S$, such that

$$
L_{\mathcal{D}}(g) \leqslant \min _{h \in \mathcal{H}} L_{\mathcal{D}}(h)+\varepsilon,
$$

where the true error function $L_{\mathcal{D}}$ is given by $L_{\mathcal{D}}(h)=\mathbb{E}_{z \sim \mathcal{D}}[\ell(h, z)]$.

### 28.4 Learning via Uniform Convergence

Definition 28.4.1 ( $\varepsilon$-representative sample). Let $\mathcal{Z}$ be a domain with distribution $\mathcal{D}$, and associated hypothesis class $\mathcal{H}$ and loss function $\ell$. We say that a sample $S$ is $\varepsilon$-representative if for every predictor $h \in \mathcal{H}$ we have

$$
\left|L_{S}(h)-L_{\mathcal{D}}(h)\right| \leqslant \varepsilon .
$$

Lemma 28.4.2. Consider a context $(\mathcal{Z}, \mathcal{D}, \mathcal{H}, \ell)$, and an $(\varepsilon / 2)$-representative sample $S$. Any output of the ERM learner

$$
\operatorname{ERM}_{\mathcal{H}}(S)=h_{S} \in \underset{h \in \mathcal{H}}{\arg \min } L_{S}(h)
$$

will be such that

$$
L_{\mathcal{D}}\left(h_{S}\right) \leqslant \min _{h \in \mathcal{H}} L_{\mathcal{D}}(h)+\varepsilon .
$$

Proof. Let $h \in \mathcal{H}$ be any predictor, then $\left|L_{S}(h)-L_{\mathcal{D}}(h)\right| \leqslant \varepsilon / 2$ therefore

$$
L_{\mathcal{D}}(h) \leqslant L_{S}(h)+\varepsilon / 2 \text { and } L_{S}(h) \leqslant L_{\mathcal{D}}(h)+\varepsilon / 2
$$

In particular, since $h_{S}$ minimises the empirical error we have

$$
\begin{aligned}
L_{\mathcal{D}}\left(h_{S}\right) & \leqslant L_{S}\left(h_{S}\right)+\varepsilon / 2 \\
& \leqslant L_{S}(h)+\varepsilon / 2 \\
& \leqslant\left(L_{\mathcal{D}}(h)+\varepsilon / 2\right)+\varepsilon / 2 \\
& =L_{\mathcal{D}}(h)+\varepsilon
\end{aligned}
$$

which proves the proposition.
Corollary 28.4.3. The $\mathrm{ERM}_{\mathcal{H}}$ learner is an agnostic PAC learner if with a probability of at least $1-\delta$ over the random choice of a sample $S$, the training set $S$ is $\varepsilon$-representative.

Definition 28.4.4 (Uniform convergence property). A hypothesis class $\mathcal{H}$ is said to satisfy the uniform convergence property-with respect to a domain $\mathcal{Z}$ and a loss function $\ell$-if there exists a minimal function

$$
m_{\mathcal{H}}^{\mathrm{UC}}:(0,1)^{2} \longrightarrow \mathbf{N}
$$

such that: for every distribution $\mathcal{D}$ over $\mathcal{Z}$, and parameters $\varepsilon, \delta \in(0,1)$, if $S \in \mathcal{Z}^{m}$ is an i.i.d. sample with size $m \geqslant m_{\mathcal{H}}^{\mathrm{UC}}(\varepsilon, \delta)$, then with a probability of at least $1-\delta$ the sample $S$ is $\varepsilon$-representative.

Corollary 28.4.5. If a hypothesis class $\mathcal{H}$ has the uniform convergence property-with respect to $(\mathcal{Z}, \ell)$-with a function $m_{\mathcal{H}}^{\mathrm{UC}}$ then $\mathcal{H}$ is agnostically PAC learnable with a sample complexity

$$
m_{\mathcal{H}}(\varepsilon, \delta) \leqslant m_{\mathcal{H}}^{\mathrm{UC}}(\varepsilon / 2, \delta)
$$

### 28.5 Finite Hypothesis Classes are Uniform Convergent

## Important Measure Concentration Inequalities

Proposition 28.5.1 (Strong law of large numbers). Let $\left(Z_{1}, \ldots, Z_{m}\right)$ be a sequence of i.i.d. random variables with equal mean $\mu$. The strong law of large numbers states that when $m \rightarrow \infty$ the empirical average $\bar{Z}:=\frac{1}{m} \sum_{j=1}^{m} Z_{j}$ converges to the expected value $\mu$ with a probability of 1.

Lemma 28.5.2 (Markov's inequality). Let $Z$ be a non-negative random variable. Then for any $a \geqslant 0$ we have

$$
\mathbb{P}(Z \geqslant a) \leqslant \mathbb{E}[Z] / a
$$

known as the Markov's inequality.
Proof. Since $\mathbb{P}(Z \geqslant x)$ is monotonically non-increasing as a function of $x$, for any $a \geqslant 0$ we have

$$
\begin{aligned}
\mathbb{E}[Z] & =\int_{0}^{\infty} \mathbb{P}(Z \geqslant x) \mathrm{d} x \\
& \geqslant \int_{0}^{a} \mathbb{P}(Z \geqslant x) \mathrm{d} x \\
& \geqslant \int_{0}^{a} \mathbb{P}(Z \geqslant a) \mathrm{d} x \\
& =a \mathbb{P}(Z \geqslant a)
\end{aligned}
$$

therefore $\mathbb{P}(Z \geqslant a) \leqslant \mathbb{E}[Z] / a$ as wanted.
Lemma 28.5.3. Given a random variable $Z$ with values in $[0,1]$, denote $\mu:=\mathbb{E}[Z]$. Then for any $a \in(0,1)$ we have the following two upper-bounds

$$
\begin{align*}
& \mathbb{P}(Z>1-a) \geqslant \frac{\mu-(1-a)}{a}  \tag{28.1}\\
& \mathbb{P}(Z>a) \geqslant \frac{\mu-a}{1-a} \geqslant \mu-a \tag{28.2}
\end{align*}
$$

Proof. Define a random variable $X:=1-Z$. Since $Z \in[0,1]$ then $Y$ is non-negative and has $\mathbb{E}[Y]=1-\mu$. By the Markov's inequality on $Y$ we have

$$
\mathbb{P}(Z \leqslant 1-a)=\mathbb{P}(1-Z \geqslant a)=\mathbb{P}(Y \geqslant a) \leqslant \frac{\mathbb{E}[Y]}{a}=\frac{1-\mu}{a} .
$$

Therefore

$$
\mathbb{P}(Z>1-a) \geqslant 1-\frac{1-\mu}{a}=\frac{\mu-(1-a)}{a}
$$

Lemma 28.5.4 (Hoeffding's lemma). Let $X$ be a random variable taking values in the interval $[a, b]$, and such that $\mathbb{E}[X]=0$. Then for every $\lambda>0$ we have

$$
\mathbb{E}\left[e^{\lambda X}\right] \leqslant \exp \left(\frac{\lambda^{2}(b-a)^{2}}{8}\right)
$$

Proof. Notice that the exponential map $f(x):=e^{\lambda x}$ is convex, therefore by definition it follows that for any $t \in(0,1)$ and $x \in[a, b]$ we have

$$
f(x) \leqslant t f(a)+(1-t) f(b)
$$

Letting $t:=\frac{b-x}{b-a} \in[0,1]$ we find

$$
e^{\lambda x} \leqslant \frac{b-x}{b-a} e^{\lambda a}+\frac{x-a}{b-a} e^{\lambda b} .
$$

If we now consider the expectation of $f$ with respect to the random variable $X$ we find the following inequality:

$$
\begin{align*}
\mathbb{E}\left[e^{\lambda X}\right] & \leqslant \frac{b-\mathbb{E}[X]}{b-a} e^{\lambda a}+\frac{\mathbb{E}[X]-a}{b-a} e^{\lambda b} \\
& =\frac{b}{b-a} e^{\lambda a}-\frac{a}{b-a} e^{\lambda b} \tag{28.3}
\end{align*}
$$

since the expected value of $X$ is null. Given $p:=-\frac{a}{b-a}$ define a map

$$
L(h):=-h p+\log \left(1-p+p e^{h}\right)
$$

so that, for $h_{0}:=\lambda(b-a)$ we have that $e^{L\left(h_{0}\right)}$ is exactly the right-hand side of Eq. (28.3). Using the fact that $L(0)=L^{\prime}(0)=0$ and that $L^{\prime \prime}(h) \leqslant 1 / 4$ for any $h$, we find by the Taylor expansion of $L$ about 0 that

$$
L\left(h_{0}\right) \approx L(0)+L^{\prime}(0) h_{0}+\frac{L^{\prime \prime}(h)}{2} h_{0}^{2}=\frac{L^{\prime \prime}\left(h_{0}\right)}{2} \leqslant \frac{h_{0}^{2}}{8}
$$

With this we have

$$
\mathbb{E}\left[e^{\lambda X}\right] \leqslant e^{L\left(h_{0}\right)} \leqslant e^{h_{0}^{2} / 8}=e^{\frac{\lambda^{2}(b-a)^{2}}{8}},
$$

which is the bound we wanted to prove.
Lemma 28.5.5 (Hoeffding's inequality). Let $\left(Z_{1}, \ldots, Z_{m}\right)$ be a collection of i.i.d. random variables, and let $\bar{Z}:=\frac{1}{m} \sum_{j=1}^{m} Z_{j}$. If for all $1 \leqslant j \leqslant m$ we have $\mathbb{E}\left[Z_{j}\right]=\mu$ and $\mathbb{P}\left[a \leqslant Z_{j} \leqslant b\right]=1$, then for any $\varepsilon>0$ we have an upper-bound

$$
\mathbb{P}(|\bar{Z}-\mu|>\varepsilon) \leqslant 2 \exp \left(-\frac{2 m \varepsilon^{2}}{(b-a)^{2}}\right)
$$

Proof. For each $j$, let $X_{j}:=Z_{j}-\mathbb{E}\left[Z_{j}\right]$ and define $\bar{X}:=\frac{1}{m} \sum_{j=1}^{m} X_{j}$. Since the exponential function is monotonically increasing, we have that for every pair $\lambda, \varepsilon>0$ :

$$
\mathbb{P}(\bar{X} \geqslant \varepsilon)=\mathbb{P}\left(e^{\lambda X} \geqslant e^{\lambda \varepsilon}\right) \leqslant e^{-\lambda \varepsilon} \mathbb{E}\left[e^{\lambda \bar{X}}\right],
$$

where we used the Markov's inequality Lemma 28.5.2. Since the random variables are i.i.d. we have

$$
\mathbb{E}\left[e^{\lambda \bar{X}}\right]=\mathbb{E}\left[e^{\lambda \frac{1}{m} \sum_{j} X_{j}}\right]=\mathbb{E}\left[\prod_{j=1}^{m} e^{\lambda X_{j} / m}\right]=\prod_{j=1}^{m} \mathbb{E}\left[e^{\lambda X_{j} / m}\right] .
$$

Via Lemma 28.5.4, for each index $j$ we have an upper-bound

$$
\mathbb{E}\left[e^{\lambda X_{j} / m}\right] \leqslant \exp \left(\frac{\lambda^{2}(b-a)^{2}}{8 m^{2}}\right)
$$

With this bound in hands we obtain

$$
\mathbb{E}\left[e^{\lambda \bar{X}}\right]=\prod_{j=1}^{m} \exp \left(\frac{\lambda^{2}(b-a)^{2}}{8 m^{2}}\right)=\exp \left(\frac{\lambda^{2}(b-a)^{2}}{8 m}\right)
$$

Therefore it follows that

$$
\mathbb{P}(\bar{X} \geqslant \varepsilon) \leqslant \exp \left(-\lambda \varepsilon+\frac{\lambda^{2}(b-a)^{2}}{8 m}\right) .
$$

If we let $\lambda:=\frac{4 m \varepsilon}{(b-a)}^{2}$ we obtain

$$
\mathbb{P}(\bar{X} \geqslant \varepsilon) \leqslant \exp \left(-\frac{2 m \varepsilon}{(b-a)^{2}}\right) .
$$

If we now consider the case for the random variable $-\bar{X}$ we shall find

$$
\mathbb{P}(\bar{X} \leqslant-\varepsilon) \leqslant \exp \left(-\frac{2 m \varepsilon^{2}}{(b-a)^{2}}\right) .
$$

Using the union bound $\mathbb{P}(|\bar{X}| \geqslant \varepsilon) \leqslant \mathbb{P}(\bar{X} \geqslant \varepsilon)+\mathbb{P}(\bar{X} \leqslant-\varepsilon)$ we obtain the wanted inequality.

### 28.6 Uniform Convergence for Finite Hypothesis Classes

Proposition 28.6.1. Let $\mathcal{H}$ be a finite hypothesis class with respect to a domain $\mathcal{Z}$, and let $\ell: \mathcal{H} \times \mathcal{Z} \rightarrow[0,1]$ be a loss function. Then $\mathcal{H}$ has the uniform convergence property with a sample complexity of

$$
m_{\mathcal{H}}^{\mathrm{UC}}(\varepsilon, \delta) \leqslant\left\lceil\frac{\log (2|\mathcal{H}| / \delta)}{2 \varepsilon^{2}}\right\rceil .
$$

Moreover, $\mathcal{H}$ is agnostically PAC learnable with respect to the ERM algorithm with a sample complexity

$$
m_{\mathcal{H}}(\varepsilon) \leqslant m_{\mathcal{H}}^{\mathrm{UC}}(\varepsilon / 2, \delta) \leqslant\left\lceil\frac{2 \log (2|\mathcal{H}| / \delta)}{\varepsilon^{2}}\right\rceil
$$

Proof. Fix parameters $\varepsilon, \delta \in(0,1)$. Lets consider the set

$$
\mathcal{S}_{\text {bad }}:=\left\{S: \exists h \in \mathcal{H},\left|L_{S}(h)-L_{\mathcal{D}}(h)\right|>\varepsilon\right\}=\bigcup_{h \in \mathcal{H}}\left\{S:\left|L_{S}(h)-L_{\mathcal{D}}(h)\right|>\varepsilon\right\}
$$

and for the sake of brevity define the notation $\mathcal{S}_{\text {bad }}^{h}:=\left\{S:\left|L_{S}(h)-L_{\mathcal{D}}(h)\right|>\varepsilon\right\}$ for each $h \in \mathcal{H}$. Via the union bound we find

$$
\mathcal{D}^{m}\left(\mathcal{S}_{\mathrm{bad}}\right) \leqslant \sum_{h \in \mathcal{H}} \mathcal{D}^{m}\left(\mathcal{S}_{\mathrm{bad}}^{h}\right)
$$

Fix a predictor $h \in \mathcal{H}$, chosen prior to the sampling of the training set. Let $S=\left(z_{j}, f\left(z_{j}\right)\right)_{j=1}^{m}$ be an i.i.d. sampled training set. Since each $z_{j}$ is sampled i.i.d. it follows that $\mathbb{E}\left[\ell\left(h, z_{j}\right)\right]=L_{\mathcal{D}}(h)$ from the definition of the true error. By the linearity of the expected value, we also have

$$
\mathbb{E}\left[L_{S}(h)\right]=\mathbb{E}\left[\frac{1}{m} \sum_{j=1}^{m} \ell\left(h, z_{j}\right)\right]=\frac{1}{m} \sum_{j=1}^{m} \mathbb{E}\left[\ell\left(h, z_{j}\right)\right]=L_{\mathcal{D}}(h) .
$$

It follows that $\left|L_{\mathcal{D}}(h)-L_{S}(h)\right|=\left|\mathbb{E}\left[L_{S}(h)\right]-L_{S}(h)\right|$ and we would wish to show that $L_{S}(h)$ is concentrated around its expected value.

Since $L_{S}(h)$ is composed of the sum of $m$ i.i.d. random variables, we can apply the law of large numbers to conclude that when $m \rightarrow \infty$ the empirical average $L_{S}(h)$ converges to its true expected value $L_{\mathcal{D}}(h)$. We would like to quantify this for finite $m$, and for that we shall make use of the Hoeffding's inequality. For that, consider the sequence of random variables $\left(\ell\left(h, z_{1}\right), \ldots, \ell\left(h, z_{m}\right)\right)$, which are sampled i.i.d. since $h$ is a fixed hypothesis and $\left(z_{1}, \ldots, z_{m}\right)$ are i.i.d. random variables. Since $\operatorname{im} \ell \subseteq[0,1]$ and that $z_{j} \in[0,1]$ for all indices $j$. We have satisfied every condition of Lemma 28.5.5 and thus we may apply the Hoeffding's inequality as follows:

$$
\mathcal{D}^{m}\left(\mathcal{S}_{\mathrm{bad}}^{h}\right)=\mathbb{P}\left(\left|L_{S}(h)-L_{\mathcal{D}}(h)\right|>\varepsilon\right) \leqslant 2 e^{-2 m \varepsilon}
$$

Therefore going back to $\mathcal{S}_{\text {bad }}$ we find

$$
\mathcal{D}^{m}\left(\mathcal{S}_{\text {bad }}\right) \leqslant \sum_{h \in \mathcal{H}} \mathcal{D}^{m}\left(\mathcal{S}_{\text {bad }}^{h}\right)=2|\mathcal{H}| e^{-2 m \varepsilon} .
$$

Therefore, by choosing an integer $m \in \mathbf{N}$

$$
m \geqslant \frac{\log (2|\mathcal{H}| / \delta)}{2 \varepsilon^{2}}
$$

we ensure that $\mathcal{D}^{m}\left(\mathcal{S}_{\text {bad }}\right) \leqslant \delta$, therefore

$$
\mathcal{D}^{m}\left(\left\{S: \forall h \in \mathcal{H},\left|L_{S}(h)-L_{\mathcal{D}}(h)\right| \leqslant \varepsilon\right\}\right) \geqslant 1-\delta .
$$

### 28.7 Universal Learners are Impossible

Theorem 28.7.1 (No-free-lunch). Let $A$ be a learning algorithm assigned for the binary classification of a domain $\mathcal{X}$ with respect to the $\{0,1\}$-loss function $\ell_{\text {class }}$. Let $m<|\mathcal{X}| / 2$ be the size of a training set. There exists a distribution $\mathcal{D}$ over the domain $\mathcal{X} \times\{0,1\}$ for which
(a) There exists a labelling map $f: \mathcal{X} \rightarrow\{0,1\}$ for which $L_{\mathcal{D}}(f)=0$.
(b) With a probability of at least $1 / 7$ over the choice of the training set $S \sim \mathcal{D}^{m}$ we have

$$
L_{\mathcal{D}}(A(S)) \geqslant \frac{1}{8} .
$$

In other words, there exists a true labelling map $f$, with respect to the distribution $\mathcal{D}$, underlying the learning problem, and the algorithm fails to output a predictor with a good approximation of $f$.

Proof. Let $C \subseteq \mathcal{X}$ be a subset of size $2 m$. We'll prove that if a learner has access to only half of $C$, then it has no information on how to correctly label the remaining half of $C$. The number of maps $C \rightarrow\{0,1\}$ is given by $T:=|\{0,1\}|^{|C|}=2^{2 m}$ : let $\left(f_{1}, \ldots, f_{T}\right)$ denote all such maps. For each $1 \leqslant j \leqslant T$ define a distribution $\mathcal{D}_{j}$ over $C \times\{0,1\}$ where

$$
\mathcal{D}_{j}(\{(x, y)\})= \begin{cases}\frac{1}{|C|}, & \text { if } y=f_{i}(x) \\ 0, & \text { otherwise }\end{cases}
$$

That is, $\mathcal{D}_{j}$ makes $f_{j}$ the true labelling function on $C$-so that $L_{\mathcal{D}_{j}}\left(f_{j}\right)=0$.
Let $K=(2 m)^{m}$ denote the number of possible sequences consisting of $m$ i.i.d. sampled instances from $C$, and let $\left(S_{1}, \ldots, S_{K}\right)$ denote all such sequences. Given a sequence $S_{j}=\left(x_{1}, \ldots, x_{m}\right)$, define $S_{j}^{i}:=\left(x_{j}, f_{j}\right)_{j=1}^{m}$ to be the labelling of $S_{j}$ by the map $f_{i}$. Given a distribution $D_{i}$, the available training sets for the algorithm $A$ are $\left(S_{1}^{i}, \ldots, S_{K}^{i}\right)$, which are i.i.d. and labelled by the same map $f_{i}$. Therefore one has

$$
\begin{equation*}
\mathbb{E}_{S \sim \mathcal{D}_{i}^{m}}\left[L_{\mathcal{D}_{i}}(A(S))\right]=\frac{1}{K} \sum_{j=1}^{K} L_{\mathcal{D}_{i}}\left(A\left(S_{j}^{i}\right)\right) \tag{28.4}
\end{equation*}
$$

Considering all $T$ pairs $\left(f_{j}, \mathcal{D}_{j}\right)$ we have the following relations:

$$
\begin{align*}
\max _{1 \leqslant i \leqslant T} \mathbb{E}_{S \sim \mathcal{D}_{i}^{m}}\left[L_{\mathcal{D}_{i}}(A(S))\right] & \geqslant \frac{1}{T} \sum_{i=i}^{T} \mathbb{E}_{S \sim \mathcal{D}_{i}^{m}}\left[L_{\mathcal{D}_{i}}(A(S))\right] \\
& =\frac{1}{T} \sum_{i=i}^{T}\left(\frac{1}{K} \sum_{j=1}^{K} L_{\mathcal{D}_{i}}\left(A\left(S_{j}^{i}\right)\right)\right) \\
& =\frac{1}{K} \sum_{j=1}^{K}\left(\frac{1}{T} \sum_{i=1}^{T} L_{\mathcal{D}_{i}}\left(A\left(S_{j}^{i}\right)\right)\right) \\
& \geqslant \min _{1 \leqslant j \leqslant K} \frac{1}{T} \sum_{i=1}^{T} L_{\mathcal{D}_{i}}\left(A\left(S_{j}^{i}\right)\right) . \tag{28.5}
\end{align*}
$$

Fix any $1 \leqslant j \leqslant K$ and let $S_{j}:=\left(x_{1}, \ldots, x_{m}\right)$ and let $\left(v_{1}, \ldots, v_{p}\right)$ be a sequence containing all instances of $C$ not appearing in $S_{j}$-which certainly has $p \geqslant m$. Hence, for any
labelling $h: C \rightarrow\{0,1\}$ and any $1 \leqslant i \leqslant T$ we have a true error

$$
\begin{aligned}
L_{\mathcal{D}_{i}}(h) & =\frac{1}{2 m} \sum_{x \in C} \ell_{\text {class }}\left(h,\left(x, f_{i}(x)\right)\right) \\
& \geqslant \frac{1}{2 m} \sum_{k=1}^{p} \ell_{\text {class }}\left(h,\left(v_{k}, f_{i}\left(v_{k}\right)\right)\right) \\
& \geqslant \frac{1}{2 p} \sum_{k=1}^{p} \ell_{\text {class }}\left(h,\left(v_{k}, f_{i}\left(v_{k}\right)\right)\right) .
\end{aligned}
$$

From this we obtain that

$$
\begin{align*}
\frac{1}{T} \sum_{i=1}^{T} L_{\mathcal{D}_{i}}\left(A\left(S_{j}^{i}\right)\right) & \geqslant \frac{1}{T} \sum_{i=1}^{T}\left(\frac{1}{2 p} \sum_{k=1}^{p} \ell_{\text {class }}\left(A\left(S_{j}^{i}\right),\left(v_{k}, f_{i}\left(v_{k}\right)\right)\right)\right) \\
& =\frac{1}{2 p} \sum_{k=1}^{p}\left(\frac{1}{T} \sum_{i=1}^{T} \ell_{\text {class }}\left(A\left(S_{j}^{i}\right),\left(v_{k}, f_{i}\left(v_{k}\right)\right)\right)\right) \\
& \geqslant \frac{1}{2} \min _{1 \leqslant k \leqslant p} \frac{1}{T} \sum_{i=1}^{T} \ell_{\text {class }}\left(A\left(S_{j}^{i}\right),\left(v_{k}, f_{i}\left(v_{k}\right)\right)\right) . \tag{28.6}
\end{align*}
$$

For the last time, fix an index $1 \leqslant k \leqslant p$. Partition $\left(f_{1}, \ldots, f_{T}\right)$ into a collection of $T / 2$ disjoint pairs of the form $\left(f_{i}, f_{i^{\prime}}\right)$ for which $f_{i}(c) \neq f_{i^{\prime}}(c)$ if and only if $c=v_{k}$. From construction, such pairs satisfy $S_{j}^{i}=S_{j}^{i^{\prime}}$ for any $1 \leqslant j \leqslant K$, therefore

$$
\ell_{\text {class }}\left(A\left(S_{j}^{i}\right),\left(v_{k}, f_{i}\left(v_{k}\right)\right)\right)+\ell_{\text {class }}\left(A\left(S_{j}^{i^{\prime}}\right),\left(v_{k}, f_{i^{\prime}}\left(v_{k}\right)\right)\right)=1,
$$

which in turn implies in

$$
\begin{equation*}
\frac{1}{T} \sum_{i=1}^{T} \ell_{\text {class }}\left(A\left(S_{j}^{i}\right),\left(v_{k}, f_{i}\left(v_{k}\right)\right)=\frac{1}{2}\right. \tag{28.7}
\end{equation*}
$$

Now substituting Eq. (28.7) in Eq. (28.6) results in

$$
\frac{1}{T} \sum_{i=1}^{T} L_{\mathcal{D}_{i}}\left(A\left(S_{j}^{i}\right)\right) \geqslant \frac{1}{4}
$$

If we now substitute this into Eq. (28.5) we get

$$
\max _{1 \leqslant i \leqslant T} \mathbb{E}_{S \sim \mathcal{D}_{i}^{m}}\left[L_{\mathcal{D}_{i}}(A(S))\right] \geqslant \frac{1}{4},
$$

and say this maximum is attained at an index $1 \leqslant i_{0} \leqslant T$. Denote $\left(f_{i_{0}}, \mathcal{D}_{i_{0}}\right):=(f, \mathcal{D})$ for short. Using the Lemma 28.5.3 we have

$$
\mathbb{P}\left(L_{\mathcal{D}}(A(S)) \geqslant 1 / 8\right) \geqslant \frac{\mathbb{E}_{S \sim D^{m}}\left[L_{\mathcal{D}}(A(S))\right]-1 / 8}{1-1 / 8}=\frac{1}{7}
$$

Corollary 28.7.2. Let $\mathcal{X}$ be an infinite domain and define $\mathcal{H}:=\{0,1\}^{\mathcal{X}}$ to be our hypothesis class-that is, the collection of all possible binary classifiers of $\mathcal{X}$. Then $\mathcal{H}$ is not PAC learnable.

Proof. Suppose, for the sake of contradiction, that $\mathcal{H}$ is PAC learnable. Take $\varepsilon<1 / 8$ and $\delta<1 / 7$. From assumption, there exists an algorithm $A$ and a number $m \in \mathbf{N}$, depending on the choice of $\varepsilon$ and $\delta$, such that for any distribution $\mathcal{D}$ over $\mathcal{X} \times\{0,1\}$ and labelling function $f: \mathcal{X} \rightarrow\{0,1\}$ with $L_{\mathcal{D}}(f)=0$-then, with a probability of at least $1-\delta$ over a sample $S \sim \mathcal{D}^{m}$, the algorithm $A$ will output a classifier with $L_{\mathcal{D}}(A(S)) \leqslant \varepsilon$. Since $\mathcal{X}$ is infinite, then certainly its cardinality overcomes $2 m$ and by Theorem 28.7.1 the algorithm $A$ fails for some distribution $\mathcal{D}_{0}$ with a probability greater than $1 / 7$-in this case the failure of $A$ means $L_{\mathcal{D}_{0}}(A(S))>1 / 8$. This is a contradiction and therefore proves that $\mathcal{H}$ is not PAC learnable.

Definition 28.7.3 (Threshold function). Let $P$ be a proposition taking values in the set $\{$ false, true $\}$. We define a map $1_{P}:\{$ false, true $\} \rightarrow\{0,1\}$ as

$$
\mathbf{1}_{P}(x)= \begin{cases}1, & \text { if } P(x)=\text { true } \\ 0, & \text { if } P(x)=\text { false }\end{cases}
$$

A common use of such maps are threshold functions: for each $a \in \mathbf{R}$ we define a map $\mathbf{1}_{<a}$ assuming 1 if the argument is less than $a$ and 0 otherwise.

Lemma 28.7.4. Let $\mathcal{H}$ denote the infinite collection of all threshold functions as in Definition 28.7.3. Then $\mathcal{H}$ is PAC learnable with the ERM algorithm with a sample complexity of

$$
m_{\mathcal{H}}(\varepsilon, \delta) \leqslant\left\lceil\frac{\log (2 / \delta)}{\varepsilon}\right\rceil
$$

Proof. Via the realisability assumption, let $a^{\star} \in \mathbf{R}$ be such that $L_{\mathcal{D}}\left(\mathbf{1}_{a^{\star}}\right)=0$. Consider the marginal distribution $\mathcal{D}_{\mathcal{X}}$ and take points $a_{0}<a^{\star}<a_{1}$ such that ${ }^{2}$

$$
\mathbb{P}_{x \sim \mathcal{D}_{X}}\left(a_{0}<x<a^{\star}\right)=\mathbb{P}_{x \sim \mathcal{D}_{X}}\left(a^{\star}<x<a_{1}\right)=\varepsilon .
$$

Let $S \sim \mathcal{D}^{m}$ be a training set, and define constants ${ }^{3}$

$$
b_{0}:=\max _{(x, 1) \in S} x \quad \text { and } \quad b_{1}:=\min _{(x, 0) \in S} x
$$

Let $b_{S} \in \mathbf{R}$ be a threshold obtained by ERM $_{\mathcal{H}}$, giving $b_{0}<b_{S}<b_{1}$. Notice that in order to have a true error $L_{\mathcal{D}}\left(\mathbf{1}_{<b_{S}}\right) \leqslant \varepsilon$ it is sufficient to ensure that $a_{0} \leqslant b_{0}$ and $b_{1} \leqslant a_{1}$ so that $b_{S} \in\left(a_{0}, a_{1}\right)$. Notice that

$$
\begin{aligned}
\mathbb{P}_{S \sim \mathcal{D}^{m}}\left(L_{\mathcal{D}}\left(\mathbf{1}_{<b_{S}}\right)>\varepsilon\right) & \leqslant \mathbb{P}_{S \sim \mathcal{D}^{m}}\left(b_{0}<a_{0} \text { and } b_{1}>a_{1}\right) \\
& \leqslant \mathbb{P}_{S \sim \mathcal{D}^{m}}\left(b_{0}<a_{0}\right)+\mathbb{P}\left(b_{1}>a_{1}\right)
\end{aligned}
$$

[^40]Notice that in order to have the event $b_{0}<a_{0}$ one has to sample every point $(x, y) \in S$ with $x<a_{0}$, since the probability of having a sample $x \in\left(a_{0}, a^{\star}\right)$ is $\varepsilon$ it follows that

$$
\begin{aligned}
\mathbb{P}_{S \sim D^{m}}\left(b_{0}<a_{0}\right) & =\mathbb{P}_{S \sim \mathcal{D}^{m}}\left(\forall(x, y) \in S, x \notin\left(a_{0}, a^{\star}\right)\right) \\
& =(1-\varepsilon)^{m} \\
& \leqslant e^{-m \varepsilon} \\
& \leqslant e^{-\frac{\log (2 / \delta)}{\varepsilon} \varepsilon} \\
& =\delta / 2
\end{aligned}
$$

since $m>\log (2 / \delta) / \varepsilon$. Analogously, $b_{1}>a_{1}$ only if each point $(x, y)$ of $S$ has $x>a_{1}$, therefore $\mathbb{P}_{S \sim D^{m}}\left(b_{1}>a_{1}\right) \leqslant \delta / 2$. This shows that

$$
\mathbb{P}_{S \sim \mathcal{D}^{m}}\left(L_{\mathcal{D}}\left(\mathbf{1}_{<b_{S}}>\varepsilon\right)\right) \leqslant \delta
$$

Definition 28.7.5 (Restricting the hypothesis class). Let $\mathcal{H}$ be a hypothesis class of maps $\mathcal{X} \rightarrow \mathcal{Y}$ and take a finite set $C \subseteq \mathcal{X}$. We define $\left.\mathcal{H}\right|_{C}:=\left(\left.h\right|_{C}\right)_{h \in \mathcal{H}}$ to be the restriction of the hypothesis class to $C$. We say that $\mathcal{H}$ shatters $C$ if it is the case that $\left.\mathcal{H}\right|_{C}=\mathcal{y}^{C}$.

Corollary 28.7.6. Let $\mathcal{H}$ be a hypothesis class of maps $\mathcal{X} \rightarrow\{0,1\}$, and $m$ be the training set size. Assume the existence of a subset $C \subseteq \mathcal{X}$ with cardinality $2 m$ such that $\left.\mathcal{H}\right|_{C}$ shatters $C$. It follows that for any learning algorithm $A$ there exists a distribution $\mathcal{D}$ over $\mathcal{X} \times\{0,1\}$ and a predictor $h \in \mathcal{H}$ for which $L_{\mathcal{D}}(h)=0$ and having a probability of at least $1 / 7$ over the choice of samples $S \sim \mathcal{D}^{m}$ that $L_{\mathcal{D}} A(S) \geqslant 1 / 8$.

Definition 28.7.7 (VC-dimension). Let $\mathcal{H}$ be a hypothesis class. We define the VCdimension of $\mathcal{H}$ to be:

VCdim $\mathcal{H}:=\max \left\{|C|: C \subseteq \mathcal{X}\right.$ is shattered by $\left.\left.\mathcal{H}\right|_{C}\right\}$.
Theorem 28.7.8. If $\mathcal{H}$ is a class with infinite VC-dimension, then $\mathcal{H}$ is not PAC learnable.
Proof. Let $S$ be a training set with size $m$. Since VCdim $\mathcal{H}=\infty$ then there exists a set with size $2 m$ that is shattered by $\mathcal{H}$ —thus by Corollary 28.7.6 $\mathcal{H}$ isn't PAC learnable.

### 28.8 Fundamental Theorem of PAC Learning

Definition 28.8.1 (Growth function). Let $\mathcal{H}$ be a hypothesis class. We define the growth function of $\mathcal{H}$ to be the map Growth $\mathcal{H}: \mathbf{N} \rightarrow \mathbf{N}$ given by

$$
\operatorname{Growth}_{\mathcal{H}}(m):=\max \left\{\left|\mathcal{H}_{C}\right|: C \subseteq \mathcal{X} \text { with }|C|=m\right\}
$$

Corollary 28.8.2. If VCdim $\mathcal{H}=d$ then for any $m \leqslant d$ we have $\operatorname{Growth}_{\mathcal{H}}(m)=2^{m}$.

Lemma 28.8.3 (Sauer-Shelah-Perles). Let $\mathcal{H}$ be a hypothesis class with VCdim $\mathcal{H} \leqslant d$ finite. Then for any $m \in \mathbf{N}$ one has

$$
\operatorname{Growth}_{\mathcal{H}}(m) \leqslant \sum_{j=0}^{d}\binom{m}{j} .
$$

In particular, if $m>d+1$ then our bound becomes

$$
\operatorname{Growth}_{\mathcal{H}}(m) \leqslant\left(\frac{e m}{d}\right)^{d}
$$

Theorem 28.8.4 (Fundamental theorem of statistical learning). Let $\mathcal{H}$ be a hypothesis class of maps $\mathcal{X} \rightarrow\{0,1\}$ together with the binary loss function. The following conditions are equivalent:
(a) The collection $\mathcal{H}$ has the uniform convergence property.
(b) Any ERM rule succeeds at PAC learning $\mathcal{H}$ agnostically.
(c) The collection $\mathcal{H}$ is agnostic PAC learnable.
(d) The collection $\mathcal{H}$ is PAC learnable.
(e) Any ERM rule succeeds at PAC learning $\mathcal{H}$.
(f) The collection $\mathcal{H}$ has a finite VC-dimension.

Theorem 28.8.5 (Sample complexity estimation). Let $\mathcal{H}$ be a hypothesis class of maps $\mathcal{X} \rightarrow\{0,1\}$ together with the binary loss function, and having a finite VCdim $\mathcal{H}=d$. There exists constants $c_{1}, c_{2} \in \mathbf{R}$ such that:
(a) $\mathcal{H}$ has the uniform convergence property and sample complexity:

$$
c_{1} \frac{d+\log (1 / \delta)}{\varepsilon^{2}} \leqslant m_{\mathcal{H}}^{\mathrm{UC}}(\varepsilon, \delta) \leqslant c_{2} \frac{d+\log (1 / \delta)}{\varepsilon^{2}}
$$

(b) $\mathcal{H}$ is agnostic PAC learnable and has sample complexity:

$$
c_{1} \frac{d+\log (1 / \delta)}{\varepsilon^{2}} \leqslant m_{\mathcal{H}}^{\mathrm{UC}}(\varepsilon, \delta) \leqslant c_{2} \frac{d+\log (1 / \delta)}{\varepsilon^{2}}
$$

(c) $\mathcal{H}$ is PAC learnable with sample complexity of:

$$
c_{1} \frac{d+\log (1 / \delta)}{\varepsilon} \leqslant m_{\mathcal{H}}(\varepsilon, \delta) \leqslant c_{2} \frac{d \log (1 / \varepsilon)+\log (1 / \delta)}{\varepsilon}
$$

### 28.9 Non-Uniform Learnability

Definition 28.9.1 (Competitive hypothesis). Let $\mathcal{H}$ be a hypothesis class, and let $h, h^{\prime} \in$ $\mathcal{H}$ be a pair of predictors, and consider parameters $\varepsilon, \delta \in(0,1)$. We say that $h$ is $(\varepsilon, \delta)$-competitive with respect to $h^{\prime}$ if with a probability of at least $1-\delta$ we have

$$
L_{\mathcal{D}}(h) \leqslant L_{\mathcal{D}}\left(h^{\prime}\right)+\varepsilon .
$$

Definition 28.9.2. We say that a hypothesis class $\mathcal{H}$ is non-uniformly learnable if there exists an algorithm $A$ and a map

$$
m_{\mathcal{H}}^{\mathrm{NUL}}:(0,1)^{2} \times \mathcal{H} \longrightarrow \mathbf{N}
$$

such that, for every triple $(\varepsilon, \delta, h) \in(0,1)^{2} \times \mathcal{H}$ the resulting predictor $A(S)$ is $(\varepsilon, \delta)$ competitive with respect to $h$. That is, with a probability of at least $1-\delta$ over the choice of sample set $S \sim \mathcal{D}^{m}$ we have

$$
L_{\mathcal{D}} A(S) \leqslant L_{\mathcal{D}}(h)+\varepsilon .
$$

Lemma 28.9.3. Let $\mathcal{H}=\bigcup_{n \in \mathrm{~N}} \mathcal{H}_{n}$ be a hypothesis class. If each $\mathcal{H}_{n}$ has the uniform convergence property, then $\mathcal{H}$ is non-uniformly learnable.

Theorem 28.9.4. Let $\mathcal{H}$ be a hypothesis class of maps of the form $\mathcal{X} \rightarrow\{0,1\}$. The class $\mathcal{H}$ is non-uniformly learnable if and only if it consists of a countable union of agnostic PAC learnable hypothesis classes.

## Chapter 29

## Geometric Deep Learning

Geometric Deep Learning

## Chapter 30

## Topological Deep Learning

### 30.1 Combinatorial Complex

## First Characterisations

Definition 30.1.1 (Combinatorial complex). A combinatorial complex (CC) consists of a triple $\mathcal{X}=(V, C$, rank $)$ where we have a set of vertices $V$, a set of cells $C \subseteq 2^{V} \backslash \varnothing$ (endowed with the inclusion preorder) and a map rank: $C \rightarrow \mathbf{Z}_{\geqslant 0}$ such that:
(a) For all $v \in V$ we have $\{v\} \in C$.
(b) The map rank is order preserving with respect to inclusions: if $x, y \subseteq C$ are such that $x \subseteq y$ then $\operatorname{rank} x \leqslant \operatorname{rank} y$.

The dimension of $\mathcal{X}$ is defined to be

$$
\operatorname{dim} \mathcal{X}:=\max _{x \in C} \operatorname{rank} x
$$

A cell $x \in C$ is said to be of $\operatorname{rank} k \in \mathbf{Z}_{\geqslant 0}$ if $\operatorname{rank} x=k$, which we summarise by saying that $x$ is a $k$-cell. The $k$-skeleton of $\mathcal{X}$ is defined as

$$
\operatorname{sk}_{k} \mathcal{X}:=\{x \in C: \operatorname{rank} x \leqslant k\} .
$$

The collection of all cells of rank $k$ will be denoted by $\mathcal{X}^{k}:=\operatorname{rank}^{-1} k$.
Notation 30.1.2. Let $X=(V, C, r a n k)$ be a combinatorial complex. For the sake of clearness we shall also adopt the following notation: vertices of $\mathcal{X}$ are denoted by Vert $\mathcal{X}:=V$, while cells of $\mathcal{X}$ are $\operatorname{Cell} \mathcal{X}:=C$, and when needed we'll also write $\operatorname{rank}_{X}=$ rank.

Example 30.1.3. The particular class of combinatorial complexes whose 1-cells (edges) are composed of exactly 2 vertices are called graph-based combinatorial complexes.

Definition 30.1.4 (Morphism of combinatorial complexes). Let $\mathcal{X}$ and $\mathcal{Y}$ be combinatorial complexes. We define a CC-morphism $\phi: \mathcal{X} \rightarrow \boldsymbol{y}$ to be a map $\phi$ : Cell $\mathcal{X} \rightarrow \operatorname{Cell} \boldsymbol{y}$ satisfying:
(a) Given a pair $x, x^{\prime} \in \operatorname{Cell} \mathcal{X}$ such that $x \subseteq x^{\prime}$, then $\phi x \subseteq \phi x^{\prime}$-that is, $\phi$ is order preserving.
(b) For any $x \in \operatorname{Cell} X$ we have $\operatorname{rank}_{x} x \geqslant \operatorname{rank}_{y}(\phi x)$-that is, for any $k \in \mathbf{Z}_{\geqslant 0}$ we have $\phi\left(\mathrm{sk}_{k} \mathcal{X}\right) \subseteq \mathrm{sk}_{k} \mathcal{Y}$.

Definition 30.1.5. A CC-morphism $\phi: \mathcal{X} \rightarrow \mathcal{Y}$ is said to be a CC-embedding if $\phi$ is an injective set function on the cells, and $\operatorname{rank}_{\mathcal{X}} x=\operatorname{rank}_{y}(\phi x)$ for every cell $x \in$ Cell $\mathcal{X}$.

Definition 30.1.6 (Sub-combinatorial complex). Let $\mathcal{X}$ be a combinatorial complex. We say that $\boldsymbol{Y}$ is a sub-combinatorial complex (or sub-CC) if there exists a CC-embedding $\boldsymbol{y} \hookrightarrow \mathcal{X}$. We shall denote by $\operatorname{Sub} \mathcal{X}$ the collection of all sub-combinatorial complexes of $\mathcal{X}$.

Example 30.1.7. For any subset $V \subseteq$ Vert $\mathcal{X}$, we can induce a sub-CC given by

$$
\mathcal{X}_{V}:=\left(V, C,\left.\operatorname{rank}_{X}\right|_{C}\right)
$$

where $C:=\{x \in \operatorname{Cell} \mathcal{X}: x \subseteq V\}$. In particular, for $k \in \mathbf{Z}_{\geqslant 0}$, the skeleton $\operatorname{sk}_{k} \mathcal{X}$ is a sub-CC of $\mathcal{X}$. A cell $x \in \operatorname{Cell} \mathcal{X}$ also induces a sub-CC whose vertices are the nodes contained in $x$ and whose cells are $\{y \in \operatorname{Cell} \mathcal{X}: y \subseteq x\}$-we shall abuse the notation and say that $x$ is a sub-CC of $\mathcal{X}$.

## Neighbourhood Functions

Definition 30.1.8 (Neighbourhood functions \& matrices). A neighbourhood function on a combinatorial complex $\mathcal{X}$ is a map $N: \operatorname{Sub} \mathcal{X} \rightarrow 2^{\text {Vert } \mathcal{X}} \backslash \varnothing$, assigning to each sub-CC of $\mathcal{X}$ a non-empty subset of nodes of $\mathcal{X}$.

Let $C=\left\{c_{1}, \ldots, c_{n}\right\}$ and $C^{\prime}=\left\{c_{1}^{\prime}, \ldots, c_{m}^{\prime}\right\}$ be two finite subsets of the cells of $\mathcal{X}$ for which $N c_{j} \subseteq C^{\prime}$ for any $1 \leqslant j \leqslant n$. We define the neighbourhood matrix of $N$ with respect to $C$ and $C^{\prime}$ to be the $m \times n$ binary matrix Mat $N=\left[a_{i j}\right]$ given by

$$
a_{i j}= \begin{cases}1, & \text { if } c_{i}^{\prime} \in N c_{j} \\ 0, & \text { otherwise }\end{cases}
$$

## Important Examples of Neighbourhood Functions

Now we shall define some of the most important neighbourhood functions for applications in topological neural networks.

Definition 30.1.9 (Down \& up incidence neighbourhood function). Let $\mathcal{X}$ be a combinatorial complex. Given a pair of distinct cells $x, y \in \operatorname{Cell} \mathcal{X}$, we say that $x$ and $y$ are incident if either $x \subsetneq y$ or $y \subsetneq x$. We define the following:
(a) The down-incidence neighbourhood function $N_{\searrow}$ : Sub $\mathcal{X} \rightarrow 2^{\operatorname{Vert} X} \backslash \varnothing$ is the map

$$
N_{\nearrow} x:=\{y \in \operatorname{Cell} \mathcal{X}: y \subsetneq x\}
$$

(b) The up-incidence neighbourhood function $N_{\nearrow}: \operatorname{Sub} \mathcal{X} \rightarrow 2^{\text {Vert } \mathcal{X}} \backslash \varnothing$ is the map

$$
N_{\nearrow} x:=\{y \in \operatorname{Cell} \mathcal{X}: x \subsetneq y\} .
$$

(c) Let $k \in \mathbf{Z}_{\geqslant 0}$ be any non-negative integer. We define the $k$-down incidence neighbourhood function $N_{\searrow, k}: \operatorname{Sub} \mathcal{X} \rightarrow 2^{\text {Vert } \mathcal{X}} \backslash \varnothing$ to be the map

$$
N_{\searrow, k} x:=\{y \in \operatorname{Cell} \mathcal{X}: y \subsetneq x \text { such that } \operatorname{rank} y=\operatorname{rank}(x)-k\} .
$$

(d) Let $k \in \mathbf{Z}_{\geqslant 0}$ be any non-negative integer. We define the $k$-up incidence neighbourhood function $N_{\searrow, k}: \operatorname{Sub} \mathcal{X} \rightarrow 2^{\text {Vert } \mathcal{X}} \backslash \varnothing$ to be the map

$$
N_{\nearrow, k} x:=\{y \in \operatorname{Cell} \mathcal{X}: x \subsetneq y \text { such that } \operatorname{rank} y=\operatorname{rank}(x)+k\} .
$$

Corollary 30.1.10. If $\mathcal{X}$ is a CC, then $N_{\searrow} x=\bigcup_{k \geqslant 0} N_{\searrow, k} x$ and $N_{\nearrow} x=\bigcup_{k \geqslant 0} N_{\nearrow, k} x$ for every $x \in \operatorname{Cell} \mathcal{X}$.

Definition 30.1.11. Given a CC $\mathcal{X}$ we define, for each cell $x \in \operatorname{Cell} \mathcal{X}$ the following:
(a) The faces of $x$ is defined to be $N_{\searrow, 1} x$.
(b) The cofaces of $x$ is defined to be $N_{\nearrow, 1} x$.

Definition 30.1.12 (Incidence matrix). Let $\mathcal{X}$ be a CC with finitely many vertices and take non-negative integers $0 \leqslant r<k \leqslant \operatorname{dim} \mathcal{X}$. We define the $(r, k)$-incidence matrix $B_{r, k}=\left[b_{i j}\right]$ between $\mathcal{X}^{r}=\left\{x_{1}, \ldots, x_{m}\right\}$ and $\mathcal{X}^{k}=\left\{x_{1}^{\prime}, \ldots, x_{n}^{\prime}\right\}$ to be the binary $m \times n$ matrix such that

$$
b_{i j}= \begin{cases}1, & \text { if } x_{i} \text { is incident to } x_{j}^{\prime} \\ 0, & \text { otherwise }\end{cases}
$$

Definition 30.1.13 ((Co)adjacency neighbourhood function). Let $\mathcal{X}$ be a CC. We shall define the following neighbourhood functions on $\mathcal{X}$ :
(a) The adjacency neighbourhood function $N_{\text {ad }}$ : Cell $\mathcal{X} \rightarrow 2^{\text {Vert } X} \backslash \varnothing$ to be the map sending a cell $x \in \operatorname{Cell} \mathcal{X}$ to the collection of all $y \in \operatorname{Cell} \mathcal{X}$ having rank $y=\operatorname{rank} x$ for which there exists $z \in \operatorname{Cell} \mathcal{X}$ such that $\operatorname{rank} z>\operatorname{rank} x$ and $x, y \subsetneq z$.
(b) The coadjacency neighbourhood function $N_{\text {coad }}$ : Cell $\mathcal{X} \rightarrow 2^{\text {Vert } X} \backslash \varnothing$ to be the map sending a cell $x \in \operatorname{Cell} \mathcal{X}$ to the collection of all $y \in \operatorname{Cell} \mathcal{X}$ having rank $y=\operatorname{rank} x$ for which there exists $z \in \operatorname{Cell} \mathcal{X}$ such that $\operatorname{rank} z<\operatorname{rank} x$, and having both $z \subsetneq y$ and $z \subsetneq x$.

A cell $z \in \operatorname{Cell} \mathcal{X}$ that satisfies either of the conditions imposed by the sets $N_{\mathrm{ad}} x$ or $N_{\text {coad }} x$ is said to be a bridge cell.

## Part X

## Appendix

## Appendix A

## Calculus on Several Variables

## A. 1 Prelude

## Sequences

> Add important facts about sequences as they come up, just as a way to collect important results

## Behaviour of Maps

Before we start our journey through differential calculus on several variables, I would like to point out some really important definitions for classifying the behaviour of maps - of which we'll use extensively.

Definition A.1.1 (Ultimately). We say that a property $P$ is ultimately satisfied by a function $f$ over a filter base $\mathcal{B}$ if there exists a $B \in \mathcal{B}$ such that $P\left(\left.f\right|_{B}\right)$.

Definition A.1.2 (Little-oh). A function $f$ is said to be little-oh (or infinitesimal) of another function $g$, which we write as $f=_{\mathcal{B}} o(g)$, if there exists a function $\alpha$ such that $f(x)=\alpha(x) g(x)$ holds ultimately over $\mathcal{B}$, and $\alpha$ is infinitesimal over $\mathcal{B}$.

Definition A.1.3 (Big-oh). Given functions $f$ and $g$, we say that $f$ is big-oh of $g$, and write $f={ }_{\mathcal{B}} O(g)$, if there exists a function $\beta$ such that ultimately over $\mathcal{B}$ we have $f(x)=\beta(x) g(x)$, and $\beta$ is ultimately bounded over $\mathcal{B}$.

Definition A.1.4 (Order over base). We say that functions $f$ and $g$ have the same order over $\mathcal{B}$, and write $f \asymp_{\mathcal{B}} g$, if $f=_{\mathcal{B}} O(g)$ and $g=_{\mathcal{B}} O(f)$, or equivalently, if exists $a, b>0$ such that for some $B \in \mathcal{B}$ we have $a|g(x)| \leqslant|f(x)| \leqslant b|g(x)|$.

Definition A.1.5. Given functions $f, g$, we say that $f$ behaves asymptotically like $g$ over $\mathcal{B}$, and write $f \sim_{\mathcal{B}} g$, if there exists a function $\gamma$ such that $\lim _{\mathcal{B}} \gamma(x)=1$ and $f(x)=\gamma(x) g(x)$ ultimately over $\mathcal{B}$.

## Fundamental Inequalities

Now we take a look at some fundamental inequalities that are used in some of the proofs of Appendix A.3.

Lemma A.1.6. For $x>0$ we have

$$
\begin{gather*}
x^{\alpha}-\alpha x+\alpha-1 \leqslant 0, \text { for } 0<\alpha<1,  \tag{A.1}\\
x^{\alpha}-\alpha x+\alpha-1 \geqslant 0, \text { for } \alpha<0 \text { or } 1<\alpha . \tag{A.2}
\end{gather*}
$$

Proof. Let $f(x)=x^{\alpha}-\alpha x+\alpha-1$, then $f^{\prime}(x)=\alpha\left(x^{\alpha-1}-1\right)$. Notice that $f^{\prime}(1)=0$ and that for $\alpha \in(0,1)$ we have that for some $\delta>0, f^{\prime}(1-\delta)>0$ and $f^{\prime}(1+\delta)<0$, which shows that $x=1$ is a strict maximum. In the case where $\alpha<0$ or $\alpha>1, f^{\prime}(1-\delta)<0$ and $f^{\prime}(1+\delta)>0$, showing that $x=1$ is a strict minimum. The strictness comes from the fact that $f$ is monotone in the intervals $x \in(0,1)$ and $x>1$. Since $f(1)=0$, then for $\alpha \in(0,1)$ the function is non-positive, and for $\alpha<0$ or $\alpha>1$ the function is non-negative.
Proposition A.1.7 (Young's inequalities). Let $a, b>0$ and $p, q \notin\{0,1\}$ such that $p^{-1}+q^{-1}=1$. Then

$$
\begin{align*}
& a^{p^{-1}} b^{q^{-1}} \leqslant \frac{a}{p}+\frac{b}{q}, \text { if } p>1  \tag{A.3}\\
& a^{p^{-1}} b^{q^{-1}} \geqslant \frac{a}{p}+\frac{b}{q}, \text { if } p<1 \tag{A.4}
\end{align*}
$$

The equality of such relations hold only when $a=b$.
Proof. Let $\alpha=p^{-1}$ and set $x=\frac{a}{b}$. From Eq. (A.1) we have

$$
\begin{aligned}
0 & \geqslant\left(\frac{a}{b}\right)^{\frac{1}{p}}-\frac{1}{p} \frac{a}{b}+\frac{1}{p}-1=\left(\frac{a}{b}\right)^{\frac{1}{p}}-\frac{1}{p} \frac{a}{b}-\frac{1}{q} \\
\frac{1}{p} \frac{a}{b}+\frac{1}{q} & \geqslant\left(\frac{a}{b}\right)^{\frac{1}{p}} \\
\frac{a}{p}+\frac{b}{q} & \geqslant a^{\frac{1}{p}} b^{1-\frac{1}{p}}=a^{p^{-1}} b^{q^{-1}}
\end{aligned}
$$

Now, from Eq. (A.1) we have equivalently that

$$
\frac{a}{p}+\frac{b}{q} \geqslant a^{p^{-1}} b^{q^{-1}} .
$$

Proposition A.1.8 (Hólder's inequalities). Let $x_{j}, y_{j} \geqslant 0$ for $1 \leqslant j \leqslant n$ and $p^{-1}+q^{-1}=1$. Then

$$
\begin{gather*}
\sum_{1 \leqslant j \leqslant n} x_{j} y_{j} \leqslant\left(\sum_{1 \leqslant j \leqslant n} x_{j}^{p}\right)^{\frac{1}{p}}\left(\sum_{1 \leqslant j \leqslant n} y_{j}^{q}\right)^{\frac{1}{q}} \text { for } p>1  \tag{A.5}\\
\sum_{1 \leqslant j \leqslant n} x_{j} y_{j} \geqslant\left(\sum_{1 \leqslant j \leqslant n} x_{j}^{p}\right)^{\frac{1}{p}}\left(\sum_{1 \leqslant j \leqslant n} y_{j}^{q}\right)^{\frac{1}{q}} \text { for } p<1 \text { and } p \neq 0 . \tag{A.6}
\end{gather*}
$$

If $p<0$, then we need the strictness $x_{j}>0$ for all $1 \leqslant j \leqslant n$. Equality is obtained for the case where $\left(x_{j}^{p}\right)_{j=1}^{n}$ and $\left(y_{j}^{q}\right)_{j=1}^{n}$ are linearly dependent.

Proof. Define $x=\sum_{j=1}^{n} x_{j}>0$ and $y=\sum_{j=1}^{n} y_{j}>0$. We can use Eq. (A.3) with $a=\frac{x_{j}^{p}}{x}$ and $b=\frac{y_{j}^{q}}{y}$, for which we find that

$$
\frac{x_{j}}{x^{p^{-1}}} \frac{y_{j}}{y^{q^{-1}}} \leqslant \frac{1}{p} \frac{x_{j}^{p}}{x}+\frac{1}{q} \frac{y_{j}^{q}}{y}
$$

hence, summing such inequality over $1 \leqslant j \leqslant n$ we find

$$
\frac{\sum_{j=1}^{n} x_{j} y_{j}}{x^{p^{-1}} y^{q^{-1}}} \leqslant \frac{1}{p} \frac{\sum_{j=1}^{n} x_{j}^{p}}{x}+\frac{1}{q} \frac{\sum_{j=1}^{n} y_{j}^{p}}{y}=\frac{1}{p}+\frac{1}{q}=1
$$

and finally Eq. (A.5) is shown

$$
\sum_{j=1}^{n} x_{j} y_{j} \leqslant x^{p^{-1}} y^{q^{-1}}=\left(\sum_{j=1}^{n} x_{j}^{p}\right)^{\frac{1}{p}}\left(\sum_{j=1}^{n} y_{j}^{q}\right)^{\frac{1}{q}}
$$

The same equivalent proof can be made with Eq. (A.4) for Eq. (A.6). Since the equality of the Young's inequalities occurs only for $a=b$, we find that the linear dependence $x_{j}^{p}=\frac{x}{y} y_{j}^{q}$ implies in the equality of Hőlder's inequalities.

Proposition A.1.9 (Minkowski's inequalities). Let $x_{j}, y_{j} \geqslant 0$ for $1 \leqslant j \leqslant n$. Then

$$
\begin{gather*}
\left(\sum_{1 \leqslant j \leqslant n}\left(x_{j}+y_{j}\right)^{p}\right)^{\frac{1}{p}} \leqslant\left(\sum_{1 \leqslant j \leqslant n} x_{j}^{p}\right)^{\frac{1}{p}}+\left(\sum_{1 \leqslant j \leqslant n} y_{j}^{p}\right)^{\frac{1}{p}} \text { for } p>1  \tag{A.7}\\
\left(\sum_{1 \leqslant j \leqslant n}\left(x_{j}+y_{j}\right)^{p}\right)^{\frac{1}{p}} \geqslant\left(\sum_{1 \leqslant j \leqslant n} x_{j}^{p}\right)^{\frac{1}{p}}+\left(\sum_{1 \leqslant j \leqslant n} y_{j}^{p}\right)^{\frac{1}{p}} \text { for } p<1 \text {, and } p \neq 0 \tag{A.8}
\end{gather*}
$$

The equality occurs when $\left(x_{j}\right)_{1 \leqslant j \leqslant n}$ and $\left(y_{j}\right)_{1 \leqslant j \leqslant n}$ are linearly dependent.
Proof. Notice that

$$
\sum_{j=1}^{n}\left(x_{j}+y_{j}\right)^{p}=\sum_{j=1}^{n}\left(x_{j}+y_{j}\right)\left(x_{j}+y_{j}\right)^{p-1}=\sum_{j=1}^{n} x_{j}\left(x_{j}+y_{j}\right)^{p-1} \sum_{j=1}^{n} y_{j}\left(x_{j}+y_{j}\right)^{p-1}
$$

If $p>0$, then applying Eq. (A.5) we find (noting that $q=\frac{p}{p-1}$ )

$$
\sum_{j=1}^{n}\left(x_{j}+y_{j}\right)^{p} \leqslant\left[\left(\sum_{j=1}^{n} x_{j}^{p}\right)^{\frac{1}{p}}+\left(\sum_{j=1}^{n} y_{j}^{p}\right)^{\frac{1}{p}}\right]\left(\sum_{j=1}^{n}\left(x_{j}+y_{j}\right)^{p}\right)^{\frac{1}{q}}
$$

For $p<1$ with $p \neq 0$, from Eq. (A.6) we get

$$
\sum_{j=1}^{n}\left(x_{j}+y_{j}\right)^{p} \geqslant\left[\left(\sum_{j=1}^{n} x_{j}^{p}\right)^{\frac{1}{p}}+\left(\sum_{j=1}^{n} y_{j}^{p}\right)^{\frac{1}{p}}\right]\left(\sum_{j=1}^{n}\left(x_{j}+y_{j}\right)^{p}\right)^{\frac{1}{q}} .
$$

Now, dividing both inequalities by the term $\left(\sum_{j=1}\left(x_{j}+y_{j}\right)^{p}\right)^{\frac{1}{q}}$ we find respectively Eq. (A.7) and Eq. (A.8). The equality occurs the same as with Hőlder inequalities. \&

## Fixed Points and Banach Spaces

Definition A.1.10 (Fixed point). Let $f: X \rightarrow X$ be any map. A fixed point of $f$ is an element $x \in X$ such that $f(x)=x$.

Theorem A.1.11 (Fixed point theorem). Let $I \subseteq \mathbf{R}$ be a closed set and $f: I \rightarrow \mathbf{R}$ a function such that $f(I) \subseteq \mathbf{R}$ and for some fixed $\theta \in[0,1)$ we have, for all $x, y \in \mathbf{R}$ :

$$
|f(x)-f(y)| \leqslant \theta|x-y|
$$

Then there exists a unique fixed point $c \in I$, that is $f(c)=c$.
Proof. Let $x_{0} \in I$ and define the sequence $x_{n}:=f\left(x_{n-1}\right)$ for all $n \geqslant 1$. We first show that $\left(x_{n}\right)_{n \in \mathbf{N}}$ is Cauchy. Let $n>m \geqslant 1$, then

$$
\begin{equation*}
\left|x_{n}-x_{m}\right|=\left|\sum_{k=m}^{n} x_{k+1}-x_{k}\right| \leqslant \sum_{k=m}^{n}\left|x_{k+1}-x_{k}\right|=\sum_{k=m}^{n}\left|f\left(x_{k}\right)-f\left(x_{k-1}\right)\right| \tag{A.9}
\end{equation*}
$$

Since $\left|f\left(x_{k}\right)-f\left(x_{k-1}\right)\right| \leqslant \theta\left|x_{k}-x_{k-1}\right|$ and $f\left(x_{k}\right)=x_{k+1}$, we can make $\prod_{i=1}^{k} \mid f\left(x_{i}\right)-$ $f\left(x_{i-1}\right)\left|\leqslant \theta^{k} \prod_{i=1}^{n}\right| x_{i}-x_{i-1} \mid$ and divide both the inequality by $\prod_{i=1}^{k}\left|f\left(x_{i}\right)-f\left(x_{i-1}\right)\right|=$ $\prod_{i=1}^{k}\left|x_{i}-x_{i-1}\right|$ in order to obtain

$$
\begin{equation*}
\left|f\left(x_{k}\right)-f\left(x_{k-1}\right)\right| \leqslant \theta^{k}\left|x_{1}-x_{0}\right| \tag{A.10}
\end{equation*}
$$

Now we can substitute Eq. (A.10) in Eq. (A.9), then

$$
\left|x_{n}-x_{m}\right| \leqslant \sum_{k=m}^{n-1} \theta^{k}\left|x_{1}-x_{0}\right|
$$

Moreover, since $\theta \in[0,1)$ we have from the geometric series that $\sum_{k=0}^{\infty} \theta^{k}=\frac{1}{1-\theta}$, so we can conclude that

$$
\left|x_{n}-x_{m}\right| \leqslant \frac{\theta^{m}}{1-\theta}\left|x_{1}-x_{0}\right|
$$

and thus $\left(x_{n}\right)_{n \in \mathbf{N}}$ is indeed a Cauchy sequence.
Let $x_{n} \rightarrow c \in I$, since $I$ is closed and thus $C l I=I$. Since $|f(x)-f(y)| \leqslant \theta|x-y|$ the function is Lipschitz continuous, hence

$$
\lim _{n \rightarrow \infty} f\left(x_{n}\right)=f(c)=\lim _{n \rightarrow \infty} x_{n+1}=c
$$

and therefore $c$ is a fixed point of $f$.
For the uniqueness of the fixed point, let $c_{1}, c_{2}$ be fixed points of $f$, then $\mid f\left(c_{1}\right)-$ $f\left(c_{2}\right)\left|=\left|c_{1}-c_{2}\right| \leqslant \theta\right| c_{1}-c_{2} \mid$ and thus $(1-\theta)\left|c_{1}-c_{2}\right| \leqslant 0$, but $\theta \in[0,1)$, hence $c_{1}=c_{2}$.

Corollary A.1.12. Let $I \subseteq \mathbf{R}$ be closed and $f: I \rightarrow I$ be a differentiable function such that exists $\theta \in[0,1)$ for which $\left|f^{\prime}(x)\right| \leqslant \theta$, for all $x \in I$. Then there exists a unique fixed point of $f$.

Proof. Choose any distinct points $x, y \in I$, from the mean value theorem, there exists $x_{0} \in(x, y)$ such that $f(x)-f(y)=f^{\prime}\left(x_{0}\right)(x-y)$, then $|f(x)-f(y)| \leqslant \theta|x-y|$, which satisfies the condition of Theorem A.1.11, hence the proposition holds.

Definition A.1.13. Let $(V,\|\cdot\|)$ be a normed vector space. We say that a sequence $\left(x_{n}\right)_{n \in \mathbf{N}} \subseteq V$ is Cauchy with respect to the norm $\|\cdot\|$ if for all $\varepsilon>0$ there exists an index $N \in \mathbf{N}$ such that, for all $n, m \geqslant N$, we have $\left\|x_{n}-x_{m}\right\|<\varepsilon$.

Definition A.1.14 (Banach space). A normed vector space ( $V,\|\cdot\|)$ is a Banach space if every Cauchy sequence converges with respect to $\|\cdot\|$.

Definition A.1.15. Let $B$ be a Banach space. A subset $A \subseteq B$ is said to be closed if the limit of every convergent sequence in $A$ belongs to $A$.

Definition A.1.16 (Contraction). Let $B$ be a Banach space and $0<\theta<1$, then a map $f: B \rightarrow B$ is said to be a $\theta$-contraction if for all $v, w \in B$ we have

$$
\|f(v)-f(w)\| \leqslant \theta\|v-w\| .
$$

Theorem A.1.17 (Banach fixed point). Let $B$ be a Banach space and $A \subseteq B$ be a closed subset. Let $f: A \rightarrow A$ be a $\theta$-contraction. Then $f$ has a unique fixed point.

Proof. The proof of the Banach fixed point is merely the same analogous proof as the one developed in Theorem A.1.11.

Proposition A.1.18 (Fixed point stability). Let $A \subseteq B$ be a closed subspace of the Banach space $B$. Let $\Omega \subseteq B$ e an open subspace of $B$. Consider the collection $\left\{f_{x} \in\right.$ $B(A, A): x \in \Omega\}$ of $\theta$-contractions such that the map $x \mapsto f_{x}(y)$ is continuous - that is, $\lim _{x \rightarrow x_{0}} f_{x}(y)=f_{x_{0}}(y)$. Then the solution map $s: \Omega \rightarrow A$ defined as

$$
s(x)=y \text { if and only if } f_{x}(y)=y
$$

is continuous at $x_{0}$ - that is, $\lim _{x \rightarrow x_{0}} s(x)=s\left(x_{0}\right)$.
Proof. We know from Theorem A.1.17 that - given any $x \in \Omega$ - the fixed point (unique) solution can be obtained as the limit of a sequence recursively defined as $y_{j}=f_{x}\left(y_{j-1}\right)$ and $y_{0} \in A$ being any element. This way, consider such sequence $\left(y_{j}\right)_{j=1}^{\infty}$ but define $y_{0}=s\left(x_{0}\right)$. Notice that since $\sum_{j=1}^{n} y_{j}-y_{j-1}=y_{n}-y_{0}$, then we can write $y_{n}$ in the following form
$y_{n}=\sum_{j=1}^{n}\left(y_{j}-y_{j-1}\right)+y_{0}=\sum_{j=2}^{n}\left(f_{x}\left(y_{j-1}\right)-f_{x}\left(y_{j-2}\right)\right)+y_{0}=\sum_{j=1}^{n}\left(f_{x}^{j-1}\left(y_{1}\right)-f_{x}^{j-1}\left(y_{0}\right)\right)+y_{0}$.

Now observe that

$$
\sum_{j=1}^{n}\left\|f_{x}^{j-1}\left(y_{1}\right)-f_{x}^{j-1}\left(y_{0}\right)\right\| \leqslant \sum_{j=1}^{n} \theta^{j-1}\left\|y_{1}-y_{0}\right\|=\frac{\left\|y_{1}-y_{0}\right\|}{1-q}
$$

That is, $\left\|y_{n}-y_{0}\right\| \leqslant \frac{\left\|y_{1}-y_{0}\right\|}{1-q}$, hence - since $\left(y_{j}\right)_{j=1}^{\infty}$ converges to the fixed point of $f_{x}$, we have

$$
\left\|s(x)-s\left(x_{0}\right)\right\|=\left\|f(s(x))-f\left(s\left(x_{0}\right)\right)\right\| \leqslant \frac{\left\|y_{1}-y_{0}\right\|}{1-q}=\frac{\left\|f_{x}\left(s\left(x_{0}\right)\right)-f_{x_{0}}\left(s\left(x_{0}\right)\right)\right\|}{1-q} .
$$

On the other hand we also know that $\lim _{x \rightarrow x_{0}} f_{x}(y)=f_{x_{0}}(y)$ thus

$$
\lim _{x \rightarrow x_{0}}\left\|f_{x}\left(s\left(x_{0}\right)\right)-f_{x_{0}}\left(s\left(x_{0}\right)\right)\right\|=0
$$

This shows that $\lim _{x \rightarrow x_{0}}\left\|s(x)-s\left(x_{0}\right)\right\|=0$ and therefore

$$
\lim _{x \rightarrow x_{0}} s(x)=s\left(x_{0}\right)
$$

Definition A.1.19. Let $V$ and $W$ be Banach spaces. We define the set $B(V, W)$ as the collection of all linear maps $f: V \rightarrow W$.

## A. 2 Continuity

Remark A.2.1. This part will be mainly concerned with the euclidean metric space given by $\mathbf{R}^{n}$ and the metric

$$
d(x, y)=\sqrt{\sum_{j=1}^{n}\left(x_{j}-y_{j}\right)^{2}}
$$

where $x=\left(x_{j}\right)_{j=1}^{n}, y=\left(y_{j}\right)_{j=1}^{n} \in \mathbf{R}$.

## Compact sets in $\mathbf{R}^{n}$

Definition A.2.2. A set $K \subseteq \mathbf{R}^{n}$ is compact if for every open cover $C$ of $K$ there exists a finite subcover $\mathcal{U} \subseteq C$.

Definition A.2.3 (General closed interval). We define a closed interval I in $\mathbf{R}^{n}$ to be the set

$$
I=\left\{x \in R^{n}: x_{j} \in\left[a_{j}, b_{j}\right], 1 \leqslant j \leqslant n\right\} .
$$

where $a, b \in \mathbf{R}^{n}$ are the boundaries of the interval $I$.
Proposition A.2.4. A closed interval in $\mathbf{R}^{n}$ is compact.

Proof. Suppose for the sake of contradiction that $\mathcal{U}$ is a cover of $I$, closed interval in $\mathbf{R}^{n}$, such that $\mathcal{U}$ doesn't admit a finite subcover. Consider the set of bisections of $I$ in which for each component of the vectors $x \in I$, that is, we create the sets $I_{j}^{1}=\left\{x \in I: x_{j} \in\left[a_{j},\left(a_{j}+b_{j}\right) / 2\right]\right\}$ and $I_{j}^{2}=\left\{x \in I: x_{j} \in\left[\left(a_{j}+b_{j}\right) / 2, b_{j}\right]\right\}$ for each $1 \leqslant j \leqslant n$, generating $2^{n}$ subsets of $I$. Notice that since these sets are contained in $I$, at least one of those should not admit a finite subcover from $\mathcal{U}$, otherwise $I$ would be compact. Hence define such set as $I_{1}$. Now recursively do the same bisection process for $I_{1}$. We end up with a chain of nested intervals

$$
I \supset I_{1} \supset I_{2} \supset \ldots
$$

each of which does not admit a finite subcover from $\mathcal{U}$. Consider the interval $I_{m}=\{x \in$ $\left.\mathbf{R}^{n}: x_{j} \in\left[a_{j}^{m}, b_{j}^{m}\right], 1 \leqslant j \leqslant n\right\}$ from the nested chain. Notice that for each $1 \leqslant j \leqslant n$ we have that the coordinate closed intervals form again a chain of nested intervals

$$
\left[a_{j}, b_{j}\right] \supseteq\left[a_{j}^{1}, b_{j}^{1}\right] \supseteq\left[a_{j}^{2}, b_{j}^{2}\right] \supseteq \ldots
$$

hence $\lim _{t \rightarrow \infty} d\left(a_{j}^{t}, b_{j}^{t}\right)=0$. Since they form a nested chain, their intersection is nonempty and therefore there exists a point $\zeta_{j} \in\left[a_{j}^{m}, b_{j}^{m}\right]$ common to all such intervals. In doing so for $1 \leqslant j \leqslant n$ we find a point $\zeta=\left(\zeta_{j}\right)_{j=1}^{n} \in \mathbf{R}^{n}$ such that $\zeta \in I_{i}$ for all $i \geqslant 1$ and $\zeta \in I$. From the last assertion one sees that there must exist $U \in \mathcal{U}$ with $\zeta \in U$, hence exists $\varepsilon>0$ such that $B_{\zeta}(\varepsilon) \subseteq U$. Now, since $\lim _{t \rightarrow \infty} d\left(a_{j}^{t}, b_{j}^{t}\right)=0$, it must be true that there exists $M>0$ such that for all $m>M$ we have $I_{m} \subseteq B_{\zeta}(\varepsilon) \subseteq U$, which clearly covers finitely $I_{m}$. This contradicts the assumption that all $I_{m}$ couldn't be finitely covered by a subcover of $\mathcal{U}$. This shows that we cannot pick a subset of $I$ with such property, implying that $I$ itself should be compact.

Proposition A.2.5. Let $K$ be a compact set of $\mathbf{R}^{n}$, then
(a) $K$ is closed in $\mathbf{R}^{n}$.
(b) Any closed subset of $K$ is compact.

Proof. (a) Let $y \in \mathbf{R}^{n}$ be any limit point of $K$. Suppose that $y \notin K$. For each point in $K$, say $x$, denote $U_{x}$ a neighbourhood. Consider the collection $\mathcal{U}=\left\{U_{x}: x \in K\right\}$, which covers the set $K$. Since $K$ is said to be compact, there exists a finite subcover $\mathcal{U}^{\prime}=\left\{U_{x_{1}}, \ldots, U_{x_{m}}\right\} \subseteq \mathcal{U}$. From the hypothesis $y$ does not belong to $K$, we can find a neighbourhood $V_{j}$ of $y$ for $1 \leqslant j \leqslant m$ such that $U_{x_{j}} \cap V_{j}=\varnothing$. Consider now the neighbourhood $V=\bigcap_{1 \leqslant j \leqslant m} V_{j}$ of $y$. Since $K=\bigcup \mathcal{U}^{\prime}$, we have $K \cap V=\varnothing$, therefore $y$ cannot be a limit point of $K$, which is a contradiction. This implies that $y \in K$, if not, problematic neighbourhoods $V_{j}$ can be chosen.
(b) Let $C \subseteq \mathbf{R}^{n}$ be a closed set and $C \subseteq K$. Let $\mathcal{G}$ be an open cover of $C$ in $\mathbf{R}^{n}$. Notice that $\mathcal{U}=\mathcal{G} \cup\left(\mathbf{R}^{n} \backslash C\right)$ is an open cover of $\mathbf{R}^{n}$, which clearly covers $K$. Therefore there exists a finite subcover of $K, \mathcal{U}^{\prime} \subseteq \mathcal{U}$, but since $C \subseteq K$, then $\mathcal{U}^{\prime}$ also covers $C$. Since $\left(\mathbf{R}^{n} \backslash C\right) \cap C=\varnothing$, then $\mathcal{U}^{\prime} \backslash\left\{\mathbf{R}^{n} \backslash C\right\}$ is a finite subcover of $C$ from $\mathcal{G}$, therefore $C$ is a compact set.

Definition A.2.6. The diameter of a set $A \subseteq \mathbf{R}^{n}$ is defined to be

$$
d(A)=\sup _{x, y \in A} d(x, y)
$$

Definition A.2.7. A set $A \subseteq \mathbf{R}^{n}$ is said to be bounded if $d(A)$ is finite.
Proposition A.2.8. If $K \subseteq \mathbf{R}^{n}$ is compact, then $K$ is also bounded in $\mathbf{R}^{n}$.
Proof. Let $\mathcal{B}$ be the collection of all open balls around a given point $x \in \mathbf{R}^{n}$, the set $\mathcal{B}$ covers $\mathbf{R}^{n}$ and therefore also covers $K$. Notice that since $K$ is compact, there exists a finite number of open balls $\mathcal{B}^{\prime} \subseteq \mathcal{B}$ that cover $K$, hence the distance between any elements of $K$ must be finite.

Theorem A.2.9 (Heine-Borel). Let $K \subseteq \mathbf{R}^{n}$ be any set. The following statements are equivalent:
(a) $K$ is closed and bounded.
(b) $K$ is compact.

Proof. Notice that the implication $(\mathrm{b}) \Rightarrow(\mathrm{a})$ is already proven by the last two propositions (Proposition A.2.5 and Proposition A.2.8). Suppose now that $K$ is a closed and bounded set. Since $K$ is bounded, there exists a closed interval $I \supset K$. Since $I$ is compact (Proposition A.2.4) and $K$ is closed we find that $K$ itself is compact (Proposition A.2.5).

## Limits

Remark A.2.10. In this subsection we shall denote a general set as $X$.
Definition A.2.11 (Limit). Let $f: X \rightarrow \mathbf{R}^{n}$ be a map. We say that $x \in \mathbf{R}^{n}$ is the limit of $f$ over a filter base $\mathcal{B} \subseteq 2^{X}$ if for every neighbourhood $V$ of $x$ there exists $B \in \mathcal{B}$ for which $f(B) \subseteq V$.

Definition A.2.12 (Bounded). A map $f: X \rightarrow \mathbf{R}^{n}$ is said to be bounded if $f(X) \subseteq \mathbf{R}^{n}$ is bounded.

Definition A.2.13 (Ultimately bounded). Given a filter base $\mathcal{B} \subseteq 2^{X}$, a map $f: X \rightarrow \mathbf{R}^{n}$ is ultimately bounded over the base $\mathcal{B}$ if there exists $B \in \mathcal{B}$ for which $f$ is bounded.

Proposition A.2.14 (Unique limit). A map can have at most one limit over a filter base.
Proof. Let $f: X \rightarrow \mathbf{R}^{n}$ be a map and suppose that $\lim _{\mathcal{B}} f(x)=a$ and $\lim _{\mathcal{B}} f(x)=b$, where $a \neq b$. Since they are distinct points, there must exist neighbourhoods $V_{a}$ and $V_{b}$ for which $V_{a} \cap V_{b}=\varnothing$. Now, remember from the definition that there must exist $B_{a}, B_{b} \in \mathcal{B}$ such that $f\left(B_{a}\right) \subseteq V_{a}$ and $f\left(B_{b}\right) \subseteq V_{b}$. From the downward direction property of filter bases, there exists $B \subseteq B_{a} \cap B_{b}$ in $\mathcal{B}$. Since $\varnothing \notin \mathcal{B}$, then $B \neq \varnothing$, hence $f(B) \subseteq V_{a} \cap V_{b}$ is non-empty, contradicting the assumption that there could be chosen non-intersecting neighbourhoods of $a$ and $b$, which implies that $a=b$ in $\mathbf{R}^{n}$.

Proposition A.2.15. If a map has a limit over a given filter base, then the map is ultimately bounded over that filter base.

Proof. Let $f: X \rightarrow \mathbf{R}^{n}$ be a map and $\mathcal{B} \subseteq 2^{X}$ be a filter base. Assume that $\lim _{\mathcal{B}} f(x)=$ $\ell \in \mathbf{R}^{n}$. For the sake of contradiction, suppose that $f$ is not ultimately bounded over $\mathcal{B}$. Let $B_{\ell}(\varepsilon)$ be an open ball centred at $\ell$ with radius $\varepsilon>0$. From the definition of a limit over a filter, there exists $B \in \mathcal{B}$ for which $f(B) \subseteq B_{\ell}(\varepsilon)$, but $\sup _{x, y \in B_{\ell}(\varepsilon)} d(x, y)=2 \varepsilon$, which contradicts the fact that $d(f(B))$ is not bounded in $\mathbf{R}$. Therefore, $f$ needs to be ultimately bounded over $\mathcal{B}$.

Corollary A.2.16. Let $f: X \rightarrow \mathbf{R}^{n}$ and $\mathcal{B}$ be a filter base over $X$. The map has a limit $y$ over $\mathcal{B}$ if and only if each of the projection functions $\pi_{j} f$ have limit $y_{j}$. That is,

$$
\lim _{\mathcal{B}} f(x)=y \Leftrightarrow \lim _{\mathcal{B}} \pi_{j} f(x)=y_{j}, \text { for } 1 \leqslant j \leqslant n
$$

Definition A.2.17 (Cauchy sequence). A sequence $\left(x_{j}\right)_{j \in \mathbf{N}}$ of points in $\mathbf{R}^{n}$ is a Cauchy sequence if for every $\varepsilon>0$ there exists an index $N \in \mathbf{N}$ for which $d\left(x_{i}, x_{j}\right)<\varepsilon$ for all $i, j>N$.

Proposition A.2.18. A sequence $\left(x_{j}\right)_{j \in \mathbf{N}}$ of points in $\mathbf{R}^{n}$ is Cauchy if and only if $\left(x_{j}^{i}\right)_{j \in \mathbf{N}}$ is Cauchy for all $1 \leqslant i \leqslant n$, where $x_{j}=\left(x_{j}^{i}\right)_{i=1}^{n}$.
Proof. Notice that since

$$
d\left(x_{j}^{i}, x_{k}^{i}\right) \leqslant d\left(x_{j}, x_{k}\right) \leqslant \sqrt{n} \max _{1 \leqslant i \leqslant n} d\left(x_{j}^{i}, x_{k}^{i}\right)
$$

then, if $\left(x_{j}\right)_{j \in \mathbf{N}}$ is Cauchy, we have that the inequality $d\left(x_{j}^{i}, x_{k}^{k}\right) \leqslant d\left(x_{j}, x_{k}\right)$ implies that $\left(x_{j}^{i}\right)_{j \in \mathbf{N}}$ is Cauchy for each $1 \leqslant i \leqslant n$. Now, if in turn we have that for each component the sequence $\left(x_{j}^{i}\right)_{j \in \mathbf{N}}$ is Cauchy, then since $d\left(x_{j}, x_{k}\right) \leqslant \sqrt{n} \max _{1 \leqslant i \leqslant n} d\left(x_{j}^{i}, x_{k}^{i}\right)$ we find that $\left(x_{j}\right)_{j \in \mathbf{N}}$ is Cauchy.

Proposition A.2.19. A sequence in $\mathbf{R}^{n}$ is convergent if and only if the sequence is Cauchy.

Proof. Suppose that $\left(x_{j}\right)_{j \in \mathbf{N}}$ is a convergent sequence in $\mathbf{R}^{n}$ with $x_{j} \rightarrow x$. Let $\varepsilon>0$, choose any neighbourhood $U$ of $x$ such that $d(U)=\varepsilon$. Since the sequence converges, there exists $N \in \mathbf{N}$ for which $x_{j} \in U$ for all $j \geqslant N$, that is, for all $j, k \geqslant N$ we have $d\left(x_{j}, x_{k}\right)<\varepsilon$. Hence we conclude that $\left(x_{j}\right)_{j \in \mathbf{N}}$ is Cauchy. For the opposite case, let $\left(x_{j}\right)_{j \in \mathbf{N}}$ be a Cauchy sequence. Then clearly there exists an element $x$ for which every neighbourhood contains infinitely many points of $\left(x_{j}\right)_{j \in \mathbf{N}}$.

Definition A.2.20 (Oscillation). The oscillation of $f: X \rightarrow \mathbf{R}^{n}$ on $E \subseteq X$ is given by

$$
\omega(f, E)=d(f(E))
$$

Theorem A.2.21 (Cauchy criterion for several variables). Let $f: X \rightarrow \mathbf{R}^{n}$ be a map and $\mathcal{B}$ be a filter base over $X$. The map $f$ has a limit over $\mathcal{B}$ if and only if for all $\varepsilon>0$ there exist $B \in \mathcal{B}$ such that $\omega(f, B)<\varepsilon$.

Proof. Apply the Cauchy criterion for single variable maps on each of $\pi_{j} f$ for $1 \leqslant j \leqslant n$, now, using Corollary A.2.16, we see that theorem is true for $f$.

■
Theorem A.2.22. Let $g: Y \rightarrow R^{n}$ and $f: X \rightarrow Y$ be mappings. Let filter basis $\mathcal{B}_{Y}$ on $Y$ - such that $g$ has a limit over $\mathcal{B}_{Y}$ - and $\mathcal{B}_{X}$ on $X$ such that for all $B_{Y} \in \mathcal{B}_{Y}$ there exists $B_{X} \in \mathcal{B}_{X}$ for which $f\left(B_{X}\right) \subseteq B_{Y}$. Then the composition $g f: X \rightarrow \mathbf{R}^{n}$ has a limit over $\mathcal{B}_{X}$ and we have the relation

$$
\lim _{\mathcal{B}_{X}} g f(x)=\lim _{\mathcal{B}_{Y}} g(y)
$$

Proof. Apply the property of the limit of the composition of single variable maps to each of the $\pi_{j} g$ and $\pi_{j} f$. From Corollary A. 2.16 we see that the theorem is true for $g$ and $f$.

Definition A.2.23 (Limit at infinity). Let $f: E \rightarrow \mathbf{R}^{n}$, where $E \subseteq \mathbf{R}^{m}$. The filter base that yields the limit $x \rightarrow \infty$ is given by $\mathcal{B}_{\infty}=\left\{\mathbf{R}^{m} \backslash B_{a}(r): r \in \mathbf{R}\right\}$ for any fixed point $a \in \mathbf{R}^{m}$.

Definition A.2.24 (Limit to infinity). Let $f: E \rightarrow \mathbf{R}^{n}$, where $E \subseteq \mathbf{R}^{m}$, and a filter base $\mathcal{B}$ on $E$. We say that $f(x) \rightarrow \mathcal{B} \infty$ if - given any fixed point $y \in \mathbf{R}^{n}$ - any open ball $B_{y}(r) \subseteq \mathbf{R}^{n}$ is such that there exists $B \in \mathcal{B}$ for which $f(B) \subseteq \mathbf{R}^{n} \backslash B_{y}(r)$.

## Continuity

Remark A.2.25. Throughout this subsection we shall assume that $E$ is a subset of $\mathbf{R}^{m}$.
Definition A.2.26 (Continuous). A map $f: E \rightarrow \mathbf{R}^{n}$ is said to be continuous at a point $x \in E$ if for every neighbourhood $V$ of $f(x)$ there exists a neighbourhood $U \subseteq E$ of $x$ such that $f(U) \subseteq V$.

Corollary A.2.27. A map $f: E \rightarrow \mathbf{R}^{n}$ is continuous at a point $x$ if and only if $\pi_{j} f$ is continuous at $x$ for each $1 \leqslant j \leqslant n$.

Definition A.2.28 (Path). We define a path in $\mathbf{R}^{n}$ to be a continuous map between an interval $I \subseteq \mathbf{R}$ and $\mathbf{R}^{n}$.

Definition A.2.29 (Support). We define the support of a path $\gamma: I \rightarrow \mathbf{R}^{n}$ to be the image $\gamma(I)$.

Definition A.2.30 (Oscillation at a point). Let $f: E \rightarrow \mathbf{R}^{n}$ be a map and $x \in E$. We define the oscillation of $f$ at the point $x$ as the limit

$$
\omega(f, x)=\lim _{r \rightarrow+0} \omega\left(f, B_{a}(r) \cap E\right)
$$

Proposition A.2.31 (Local properties). Let $f: E \rightarrow \mathbf{R}^{n}$ be a map.
(a) $f$ is continuous at $x \in E$ if and only if $\omega(f, x)=0$.
(b) If $f$ is continuous at a point $x \in E$, then $f$ is bounded in some neighbourhood $U_{x} \cap E$ of $x$.
(c) Let set $X \subseteq \mathbf{R}^{m}$ and $Y \in \mathbf{R}^{n}$. Let $g: Y \rightarrow \mathbf{R}^{k}$ be a continuous map at $y \in Y$. Let $f: X \rightarrow Y$ be continuous at $x \in X$ and $f(x)=y$. Then the map $g f: X \rightarrow R^{k}$ is continuous at $x$.

If the map is real valued, we also have more properties. Let $f, g: E \rightarrow \mathbf{R}$, then
(a) If $f$ is continuous at a point $\bar{x} \in E$, there exists a neighbourhood $U \cap E$ of $\bar{x}$ such that $f(x) f(\bar{x})>0$ for all $x \in U \cap E$.
(b) If $f$ and $g$ are continuous at a point $x \in E$, then for any $\alpha, \beta \in \mathbf{R}$ we have that the linear combination $\alpha f+\beta g: E \rightarrow \mathbf{R}$, their product $f \cdot g: E \rightarrow \mathbf{R}$ and - if $g(x) \neq 0$ - the quotient $\frac{f}{g}: E \rightarrow \mathbf{R}$ are all continuous at the point $x$.

Proof.

## Write proofs: local properties

Definition A.2.32 (Uniformly continuous). Let $f: X \rightarrow Y$ be a map between metric spaces. We say that $f$ is uniformly continuous on $X$ if for every $\varepsilon>0$ there exists $\delta>0$ such that, for every subset $E \subseteq X$ of diameter $d(E)<\delta$, we have an oscillation $\omega(f, E)<\varepsilon$.

Theorem A.2.33 (Heine-Cantor theorem). Let $f: X \rightarrow Y$ be a continuous map between metric spaces $X$ and $Y$. If $X$ is compact, then $f$ is uniformly continuous.

Proof. Let any $\varepsilon>0$. Since $f$ is continuous, there exists, for any $x \in X$ a $\delta_{x}>0$ for which, if $d_{X}(x, y)<\delta_{x}$, then $d_{Y}(f(x), f(y))<\varepsilon / 2$. Lets consider the open cover $\mathcal{U}$ of $X$ consisting of the neighbourhoods $U_{x}:=\left\{y \in X: d_{X}(x, y)<\frac{\delta_{x}}{2}\right\}$ for each $x \in X$. From definition, if $X$ is compact, then there exists a finite subcover $\left\{U_{x_{j}}\right\}_{j=1}^{n} \subseteq \mathcal{U}$ of $X$. Define the minimum radius of the given neighbourhoods as $\delta:=\min _{1 \leqslant j \leqslant n} \delta_{x_{j}} / 2$.

Let $x, y \in X$ be any points such that $d_{X}(x, y)<\delta$. From the finite subcover, we have that there exists $1 \leqslant j_{0} \leqslant n$ such that $x \in U_{x_{0}}$, which implies that $d_{X}\left(x, x_{j_{0}}\right)<\delta_{x_{0}} / 2$, thus $d_{Y}\left(f(x), f\left(x_{j_{0}}\right)\right)<\varepsilon / 2$, from construction. Notice however that

$$
d_{X}\left(y, x_{j_{0}}\right) \leqslant d_{X}(y, x)+d_{X}\left(x, x_{j_{0}}\right)<\delta+\delta_{x_{j_{0}}} \leqslant \delta_{x_{j_{0}}}
$$

therefore it follows that $d_{Y}\left(f\left(x_{j_{0}}\right), f(y)\right)<\varepsilon / 2$. Using again the triangle inequality we observe that

$$
d_{Y}(f(x), f(y)) \leqslant d_{Y}\left(f(x), f\left(x_{j_{0}}\right)\right)+d_{Y}\left(f\left(x_{j_{0}}\right), f(y)\right)<\varepsilon
$$

Therefore, $f$ is indeed uniformly continuous with constant $\delta$.
Definition A.2.34 (Pathwise connected). A set $X \subseteq \mathbf{R}^{n}$ is pathwise connected if for all $x, y \in E$ there exists a path $\gamma: I \rightarrow E$ with endpoints at $x$ and $y$ and support in $E$.

Definition A.2.35 (Domain). A domain in $\mathbf{R}^{n}$ is an open pathwise connected subset of $\mathbf{R}^{n}$.

Proposition A.2.36 (Global properties). The following are global properties on continuous maps of several variables. Let $K \subseteq \mathbf{R}^{m}$ be a compact set and $E \subseteq \mathbf{R}^{m}$ be pathwise connected.
(a) A continuous map $f: K \rightarrow \mathbf{R}^{n}$ is uniformly continuous.
(b) A continuous map $f: K \rightarrow \mathbf{R}^{n}$ is bounded on $K$.
(c) A continuous map $f: K \rightarrow \mathbf{R}$ assumes its maximal and minimal values at least once in $K$.
(d) Let $f: E \rightarrow \mathbf{R}$ be a continuous map and assume $f(a)=A$ and $f(b)=B$ at $a, b \in E$. For any $C \in[A, B] \subseteq \mathbf{R}$ there exists $c \in E$ such that $f(c)=C$.

Proof.

## Write proofs: global properties

(d) From the connectedness of $E$, let $\gamma:[x, y] \rightarrow E$ be a continuous path such that $\gamma(x)=a$ and $\gamma(y)=b$. Consider the composition of continuous maps $f \gamma: I \rightarrow \mathbf{R}$. Since $f \gamma(a)=A$ and $f \gamma(b)=B$, for any given $C \in[A, B]$, there exists $z \in[x, y]$ such that $f \gamma(z)=C$, hence there exists $c=\gamma(z) \in E$ for which $f(c)=C$.

Proposition A.2.37. Every linear map of the form $L: \mathbf{R}^{m} \rightarrow \mathbf{R}^{n}$ is continuous. Moreover, it is uniformly continuous.

Proof. Let $f: E \rightarrow \mathbf{R}^{n}$ be any map, then for all $1 \leqslant j \leqslant n$ we have

$$
\left\|\pi_{j} f(x)\right\|_{\mathbf{R}} \leqslant\|f(x)\|_{\mathbf{R}^{n}} \leqslant \sum_{j=1}^{n}\left\|\pi_{j} f(x)\right\|_{\mathbf{R}}
$$

In particular, for the linear map $L$ we have that - given any $x \in \mathbf{R}^{m}$

$$
\|L(x)\|_{\mathbf{R}^{n}}=\left\|\sum_{j=1}^{m} x_{j} L\left(e_{j}\right)\right\|_{\mathbf{R}^{n}} \leqslant \sum_{j=1}^{m}\left\|x_{j}\right\|_{\mathbf{R}}\left\|L\left(e_{j}\right)\right\|_{\mathbf{R}^{n}} \leqslant\|x\|_{\mathbf{R}^{m}}\left(\sum_{j=1}^{m}\left\|L\left(e_{j}\right)\right\|_{\mathbf{R}^{n}}\right)
$$

Hence $L=O(\mathrm{id})$ - where $\mathrm{id}: \mathbf{R}^{m} \rightarrow \mathbf{R}^{m}$ mapping $x \mapsto x$. It follows from this that as $x \rightarrow x_{0}$, we have $L\left(x-x_{0}\right)=L(x)-L\left(x_{0}\right) \rightarrow 0$. This shows that $L$ is continuous at any point of $\mathbf{R}^{m}$. Notice that given any $\varepsilon>0$ if $\|x-y\|_{\mathbf{R}^{m}}<\frac{\varepsilon}{\ell}$ — where $\ell:=\sum_{j=1}^{m} L\left(e_{j}\right)$ then
$\|L(x)-L(y)\|_{\mathbf{R}^{n}}=\left\|\sum_{j=1}^{m}\left(x_{j}-y_{j}\right) L\left(e_{j}\right)\right\|_{\mathbf{R}^{n}} \leqslant \sum_{j=1}^{m}\left\|L\left(e_{j}\right)\right\|_{\mathbf{R}^{n}}\left\|x_{j}-y_{j}\right\|_{\mathbf{R}} \leqslant \ell\|x-y\|_{\mathbf{R}^{m}}<\ell \frac{\varepsilon}{\ell}=\varepsilon$
where we conclude that $L$ is uniformly continuous.

## A. 3 Differentiable Maps

Remark A.3.1. Throughout this chapter, we'll denote by $\left\|\|: \mathbf{R}^{n} \rightarrow \mathbf{R}\right.$ the standard norm in $\mathbf{R}^{n}$, given by

$$
\|x\|=\sqrt{\sum_{j=1}^{n} x_{j}^{2}}
$$

Moreover, the set $E$ is a any subset of $\mathbf{R}^{m}$.
Definition A.3.2 (Little-Oh). Given maps $f: X \rightarrow \mathbf{R}^{n}$ and $g: X \rightarrow \mathbf{R}^{m}$, we say that $f$ is little-Oh of $g$ over a filter base $\mathcal{B} \subseteq 2^{X}$ — and write $f=o(g)$ - if $\|f(x)\|_{\mathbf{R}^{n}}=$ $o\left(\|g(x)\|_{\mathbf{R}^{m}}\right)$ over $\mathcal{B}$.

Proposition A.3.3. Let $L: \mathbf{R}^{m} \rightarrow \mathbf{R}^{n}$ be a linear map, then for $h \in \mathbf{R}^{m}$, as $h \rightarrow 0$ we have

$$
L(o(h))=o(h)
$$

Proof. Define $f(h)=\alpha(h) h=o(h)$ - that is $\alpha: \mathbf{R}^{n} \rightarrow \mathbf{R}$, where $\alpha(h) \rightarrow 0$ as $h \rightarrow 0$. This yields $L(f(h))=L(\alpha(h) h)=\alpha(h) \sum_{j=1}^{n} h_{j} L\left(e_{j}\right)$, therefore

$$
\begin{align*}
\|L(f(h))\|_{\mathbf{R}^{n}} & =\|\alpha(h)\|_{\mathbf{R}}\left\|_{j=1}^{n} h_{j} L\left(e_{j}\right)\right\|_{\mathbf{R}^{n}} \\
& \leqslant\|\alpha(h)\|_{\mathbf{R}} \sum_{j=1}^{n}\left\|L\left(e_{j}\right)\right\|_{\mathbf{R}^{n}}\left\|h_{j}\right\|_{\mathbf{R}} \\
& \leqslant\left(\|\alpha(h)\|_{\mathbf{R}} \sum_{j=1}^{n}\left\|L\left(e_{j}\right)\right\|_{\mathbf{R}^{n}}\right)\|h\|_{\mathbf{R}^{m}}
\end{align*}
$$

and since $\|\alpha(h)\|_{\mathbf{R}} \sum_{j=1}^{n}\left\|L\left(e_{j}\right)\right\|_{\mathbf{R}^{n}} \rightarrow 0$ as $h \rightarrow 0$, we find that $L(o(h))=o(h)$.
Definition A.3.4 (Big-Oh). Let maps $f: X \rightarrow \mathbf{R}^{n}$ and $g: X \rightarrow \mathbf{R}^{m} . f$ is said to be big-Oh of $g$ over a filter base $\mathcal{B} \subseteq 2^{X}$ — and denote by $f=O(g)$ —if $\|f(x)\|_{\mathbf{R}^{n}}=O\left(\|g(x)\|_{\mathbf{R}^{m}}\right)$ over $\mathcal{B}$.

Proposition A.3.5. Let $L: \mathbf{R}^{m} \rightarrow \mathbf{R}^{n}$ be a linear map. For any $h \in \mathbf{R}^{m}$ we have that as $h \rightarrow 0$

$$
L(h)=O(h)
$$

Proof. Consider $h=\sum_{j=1}^{m} h_{j} e_{j}$, then $L(h)=\sum_{j=1}^{m} h_{j} L\left(e_{j}\right)$. From Minkowski's inequalities (see Proposition A.1.9) we have

$$
\|L(h)\|_{\mathbf{R}^{n}}=\left\|\sum_{j=1}^{m} h_{j} L\left(e_{j}\right)\right\|_{\mathbf{R}^{n}} \leqslant \sum_{j=1}^{m}\left\|L\left(e_{j}\right)\right\|_{\mathbf{R}^{n}}\left\|h_{j}\right\|_{\mathbf{R}} \leqslant\left(\sum_{j=1}^{m}\left\|L\left(e_{j}\right)\right\|_{\mathbf{R}^{n}}\right)\|h\|_{\mathbf{R}^{m}}=M\|h\|_{\mathbf{R}^{m}}
$$

Thus, as $h \rightarrow 0$ that $L(h)=O(h)$.

Definition A.3.6 (Tangent space). Let $x \in \mathbf{R}^{n}$. We define the tangent space of $x$ denoted by $T_{x} \mathbf{R}^{n}$ - to be the $R$-vector space spanned by $x$.

Definition A.3.7 (Differentiable). A map $f: E \rightarrow \mathbf{R}^{n}$ is differentiable at a point $x \in E$ - where $x$ is a limit point of $E$ - if

$$
\begin{equation*}
f(x+t)-f(x)=L(x)(t)+\alpha(x, t) \tag{A.11}
\end{equation*}
$$

where $L(x): \mathbf{R}^{m} \rightarrow \mathbf{R}^{n}$ is a linear map - on the variable $t$ - and $\alpha(x, t)=o(t)$ as $t \rightarrow 0$. We call the linear map $L(x)$ the differential (or tangent map) of $f$ at the point $x \in E$ - we shall normally denote $L(x)$ by $\mathrm{d} f(x): T_{x} \mathbf{R}^{m} \rightarrow T_{f(x)} \mathbf{R}^{n}$. We can clarify the last equation by writing it equivalently as

$$
\begin{equation*}
f(x+t)-f(x)=\mathrm{d} f(x)(t)+o(t) . \tag{A.12}
\end{equation*}
$$

Proposition A.3.8. A map $f: E \rightarrow \mathbf{R}^{n}$ is differentiable at a limit point $x \in E$ if and only if, for all $1 \leqslant j \leqslant n$, the maps $\pi_{j} \circ f: E \rightarrow \mathbf{R}$ are differentiable at $x$.

Proof. Notice that Eq. (A.12) can be rewritten coordinate-wise - as $t \rightarrow 0$

$$
\begin{equation*}
\pi_{j} \circ f(x+t)-\pi_{j} \circ f(x)=\left(\pi_{j} \circ \mathrm{~d} f(x)\right)(t)+o(t) \tag{A.13}
\end{equation*}
$$

Assume that $\pi_{j} \circ \mathrm{~d} f(x)$ has the general form of $\left(\pi_{j} \circ \mathrm{~d} f(x)\right)(t):=\sum_{j=1}^{m} a_{x}^{j} t_{j}$, where $a_{x}^{j} \in \mathbf{R}$ are scalars dependent on $x \in E$ for all $1 \leqslant j \leqslant m$. Consider, for each $1 \leqslant j \leqslant m$, the displacements $t=t_{j} e_{j}$ so that $\|t\|_{\mathbf{R}^{m}}=\left\|t_{j}\right\|_{\mathbf{R}}$. For such displacements, we find as $t_{j} \rightarrow 0$

$$
\pi_{j} \circ f\left(x+t_{j} e_{j}\right)-\pi_{j} \circ f(x)=a_{x}^{j} t_{j}+o\left(t_{j}\right)
$$

therefore we find that each of the scalar corresponding to displacements on the $j$ th coordinate - of the $j$ th differential $\pi_{j} \circ f$ of $f$ - can be written as

$$
\begin{equation*}
a_{x}^{j}=\lim _{t_{j} \rightarrow 0} \frac{\pi_{j} \circ f\left(x+t_{j} e_{j}\right)-\pi_{j} \circ f(x)}{t_{j}} \tag{A.14}
\end{equation*}
$$

The equivalence between Eq. (A.12) and Eq. (A.13) shows that the map $f$ is differentiable at $x$ if and only if each of its coordinate decompositions are differentiable at $x$.

## Partial Derivative and Differential of Real Valued Maps

Definition A.3.9 (Partial derivative). Let $f: E \rightarrow \mathbf{R}$, where $E \subseteq \mathbf{R}^{m}$. We define the partial derivative of $f$ at the point $x \in \operatorname{Int} E$, with respect to its $j$ th coordinate, to be the real value (if existent)

$$
\partial_{j} f(x)=\lim _{t \rightarrow 0} \frac{f\left(x+t e_{j}\right)-f(x)}{t}
$$

Notice that, given a map $f: E \rightarrow \mathbf{R}^{n}$, the partial derivative of each of its coordinatewise maps $\pi_{j} \circ f$ is the real value

$$
\partial_{j}\left(\pi_{j} \circ f\right)(x)=\lim _{\mathrm{R} \ni t \rightarrow 0} \frac{\pi_{j} \circ f\left(x+t e_{j}\right)-\pi_{j} \circ f(x)}{t}
$$

Consider the $j$ th projection map $\pi_{j}: \mathbf{R}^{m} \rightarrow \mathbf{R}$. From the definition Definition A.3.7, the differential of $\pi_{j}$ at a point $x \in \mathbf{R}^{m}$ is given by $\mathrm{d} \pi_{j}(x)(t)=\pi_{j}(x+t)-\pi_{j}(x)=t_{j}$, which is independent of $x$. We can define now the following operator, that will stand as a clever notation for the differential of the projection maps.

Notation A.3.10. Given a point $x \in \mathbf{R}^{m}$, we denote the linear map $\mathrm{d} x_{j}: T_{x} \mathbf{R}^{m} \rightarrow T_{\pi_{j}(x)} \mathbf{R}$ as

$$
\mathrm{d} x_{j}(t)=t_{j} .
$$

Proposition A.3.11 (Real valued differential). Let $f: E \rightarrow \mathbf{R}$ be a differentiable map at an interior point $x \in \operatorname{Int} E$. Then $f$ has a partial derivative at $x$ for each of its variables. The differential of $f$ at the point $x$ - the map $\mathrm{d} f(x): T_{x} \mathbf{R}^{m} \rightarrow T_{f(x)} \mathbf{R}^{n}$ - is uniquely defined as the sum of the partial derivatives of $f$ at $x$ - that is, for any $t \in T_{x} \mathbf{R}^{m}$, we have

$$
\mathrm{d} f(x)(t)=\sum_{1 \leqslant j \leqslant m} \partial_{j} f(x) t_{j}
$$

In view of the introduced Notation A.3.10, we can rewrite the differential as the map

$$
\mathrm{d} f(x)=\sum_{1 \leqslant j \leqslant m} \partial_{j} f(x) \mathrm{d} x_{j} .
$$

Notation A.3.12. For the sake of brevity - given a function $f: E \rightarrow R^{n}$ - we define $f_{j}:=\pi_{j} \circ f$ for each $1 \leqslant j \leqslant n$.

## Differential of a Map $\mathbf{R}^{m} \rightarrow \mathbf{R}^{n}$

Now we can generalize the results for real valued functions $E \rightarrow \mathbf{R}$ to functions of the type $E \rightarrow \mathbf{R}^{n}$.

Corollary A.3.13 (Several variables differential). Let $f: E \rightarrow \mathbf{R}^{n}$ be differentiable at the interior point $x \in \operatorname{Int} E$. The differential of $f$ at $x, \mathrm{~d} f(x): T_{x} \mathbf{R}^{m} \rightarrow T_{f(x)} \mathbf{R}^{m}$, exists and is uniquely given by - for any given $t \in T_{x} \mathbf{R}^{m}$

$$
\mathrm{d} f(x)(t)=\left[\begin{array}{c}
\mathrm{d} f_{1}(x)(t)  \tag{A.15}\\
\vdots \\
\mathrm{d} f_{n}(x)
\end{array}\right]=\left[\begin{array}{ccc}
\sum_{j=1}^{m} \partial_{j} f_{1}(x) t_{j} \\
\vdots \\
\sum_{j=1}^{m} \partial_{j} f_{n}(x) t_{j}
\end{array}\right]=\left[\begin{array}{ccc}
\partial_{1} f_{1}(x) & \ldots & \partial_{m} f_{1}(x) \\
\vdots & \ddots & \vdots \\
\partial_{1} f_{n}(x) & \cdots & \partial_{m} f_{n}(x)
\end{array}\right]\left[\begin{array}{c}
t_{1} \\
\vdots \\
t_{m}
\end{array}\right]
$$

Definition A.3.14 (Jacobi matrix). The matrix given by the partial derivatives of the projections of $f: E \rightarrow \mathbf{R}^{n}-\left[\partial_{i} f_{j}(x)\right]_{i, j}$, with $1 \leqslant i \leqslant m$ and $1 \leqslant j \leqslant n-$ is called the Jacobi matrix $f$ at $x$.

Notation A.3.15 (Jacobi matrix). We'll denote the Jacobi matrix of a map $f: E \rightarrow \mathbf{R}^{n}$ at an interior point $x$ - where $f$ is differentiable - as $f^{\prime}(x)$.

Definition A.3.16 (Jacobian). Let $E \subseteq \mathbf{R}^{n}$ and a map $f: E \rightarrow \mathbf{R}^{n}$ - where $f$ is differentiable at an interior point $x \in \operatorname{Int} E$. The Jacobi matrix of $f$ at the point $x$ is a $n \times n$ square matrix and its determinant $\operatorname{det}\left[\partial_{i} f_{j}(x)\right]_{i, j}$ is called the Jacobian of $f$ at $x$.

## Connections Between Differentiability and Continuity

Corollary A.3.17 (Differentiable implies continuous). Let $f: E \rightarrow \mathbf{R}^{n}$ be a differentiable map at a point $x \in E$. Then $f$ is continuous at $x$.

Proof. By definition, the differential of $f$ at $x$ exists and is a linear map. By Proposition A.2.37 we find that $\mathrm{d} f(x)$ is continuous and hence $\mathrm{d} f(x)(t) \rightarrow 0$ as $t \rightarrow 0$. Then as $t \rightarrow 0$ we have $f(x+t)-f(x)=\mathrm{d} f(x)(t)+o(t) \rightarrow 0$ so $\lim _{t \rightarrow 0} f(x+t)=f(x)$ - the map is continuous at $x$.

Remark A.3.18 (Differential and partial derivatives). The Proposition A.3.11 shows that if a map is differentiable at an interior point of its domain, then the partial derivatives exist at the given point. However, the converse does not hold. For instance, consider the map $g: \mathbf{R}^{2} \rightarrow \mathbf{R}$ given by

$$
g(x, y):= \begin{cases}\frac{x y}{x^{2}+y^{2}}, & x^{2}+y^{2} \neq 0 \\ 0, & x^{2}+y^{2}=0\end{cases}
$$

Notice that $g(0, y)=g(x, 0)=0$ but $g(x, x)=\frac{1}{2}$ when $x \neq 0$, therefore $g$ has no limit as $(x, y) \rightarrow 0$. On the other hand, $g$ has partial derivatives for all points over the plane

$$
\partial_{1} g(x, y)=\frac{y\left(y^{2}-x^{2}\right)}{\left(x^{2}+y^{2}\right)^{2}} \text { and } \partial_{2} g(x, y)=\frac{x\left(x^{2}-y^{2}\right)}{\left(x^{2}+y^{2}\right)^{2}}
$$

## Laws of Differentiability

Theorem A.3.19. Let $f, g: E \rightarrow \mathbf{R}^{n}$ be differentiable maps at the point $x \in E$, then the linear combination $\lambda f+\gamma g: E \rightarrow \mathbf{R}^{n}$, where $\lambda, \gamma \in \mathbf{R}$ are scalars, is differentiable at $x$ and the differential $\mathrm{d}(\lambda f+\gamma g): T_{x} \mathbf{R}^{m} \rightarrow T_{\lambda f(x)+\gamma g(x)} \mathbf{R}^{n}$ is given by

$$
\mathrm{d}(\lambda f+\gamma g)(x)=(\lambda \mathrm{d} f+\gamma \mathrm{d} g)(x)
$$

Proof. Note that

$$
\begin{aligned}
(\lambda f+\gamma g)(x+t)-(\lambda f+\gamma g)(x) & =\lambda(f(x+t)-f(x))+\gamma(g(x+t)+g(x)) \\
& =\lambda(\mathrm{d} f(x)(t)+o(t))+\gamma(\mathrm{d} g(x)(t)+o(t)) \\
& =(\lambda \mathrm{d} f(x)+\gamma \mathrm{d} g(x))(t)+o(t)
\end{aligned}
$$

thus $\lambda f+\gamma g$ is differentiable at $x$. Taking the limit as $t \rightarrow 0$, we find

$$
\mathrm{d}(\lambda f+\gamma g)(x)(t)=\lambda \mathrm{d} f(x)(t)+\gamma \mathrm{d} g(x)(t)
$$

If we also assume that $x \in \operatorname{Int} E$, then we can rewrite the last relation as

$$
\left[\begin{array}{ccc}
\partial_{1}(\lambda f+\gamma g)_{1}(x) & \ldots & \partial_{m}(\lambda f+\gamma g)_{1}(x) \\
\vdots & \ddots & \vdots \\
\partial_{1}(\lambda f+\gamma g)_{n}(x) & \ldots & \partial_{m}(\lambda f+\gamma g)_{n}(x)
\end{array}\right]=\left[\begin{array}{ccc}
\lambda \partial_{1} f_{1}(x)+\gamma \partial_{1} g_{1}(x) & \ldots & \lambda \partial_{m} f_{1}(x)+\gamma \partial_{m} g_{1}(x) \\
\vdots & \ddots & \vdots \\
\lambda \partial_{1} f_{n}(x)+\gamma \partial_{1} g_{n}(x) & \ldots & \lambda \partial_{m} f_{n}(x)+\gamma \partial_{m} g_{n}(x)
\end{array}\right]
$$

thus the following equality is obtained

$$
\partial_{i}(\lambda f+\gamma g)_{j}(x)=\lambda \partial_{i} f_{j}(x)+\gamma \partial_{i} g_{j}(x)
$$

Theorem A.3.20. Let $h, \ell: E \rightarrow \mathbf{R}$ be differentiable maps at $x \in E$. Then
(a) The product $h \cdot \ell: E \rightarrow \mathbf{R}$ is differentiable at $x$ and the differential $\mathrm{d}(h \cdot \ell): T_{x} \mathbf{R}^{m} \rightarrow$ $T_{(h \cdot \ell)(x)} \mathbf{R}$ is given by

$$
\begin{equation*}
\mathrm{d}(h \cdot \ell)(x)=h(x) \mathrm{d} \ell(x)+\ell(x) \mathrm{d} h(x) \tag{A.16}
\end{equation*}
$$

If we also assume $x \in \operatorname{Int} E$, then the relation can be rewritten matricially as

$$
\left[\begin{array}{lll}
\partial_{1}(h \cdot \ell)(x) & \cdots & \partial_{m}(h \cdot \ell)(x)
\end{array}\right]=\left[\begin{array}{lll}
h(x) \partial_{1} \ell(x)+\ell \partial_{1} h(x) & \cdots & h(x) \partial_{m} \ell(x)
\end{array}\right]
$$

thus the following relation is obtained

$$
\begin{equation*}
\partial_{i}(h \cdot \ell)(x)=h(x) \partial_{i} \ell(x)+\ell(x) \partial_{i} h(x) \quad \text { for } 1 \leqslant i \leqslant m \tag{A.17}
\end{equation*}
$$

(b) If $\ell(x) \neq 0$, the quotient $\frac{h}{\ell}: E \rightarrow \mathbf{R}$ is differentiable at $x$ and the differential of such quotient at $x, \mathrm{~d} \frac{h}{\ell}: T_{x} \mathbf{R}^{m} \rightarrow T_{\frac{h}{\ell}(x)} \mathbf{R}$, is given by

$$
\begin{equation*}
\mathrm{d} \frac{h}{\ell}(x)=\frac{\ell(x) \mathrm{d} h(x)-h(x) \mathrm{d} \ell(x)}{\ell^{2}(x)} \tag{A.18}
\end{equation*}
$$

If we also assume $x \in \operatorname{Int} E$, then the relation can be rewritten matricially as

$$
\left[\begin{array}{lll}
\partial_{1} \frac{h}{\ell}(x) & \cdots & \partial_{m} \frac{h}{\ell}(x)
\end{array}\right]=\frac{1}{\ell^{2}(x)}\left[\ell(x) \partial_{1} h(x)-h(x) \partial_{1} \ell(x) \quad \cdots \quad \ell(x) \partial_{m} h(x)-h(x) \partial_{m} \ell(x)\right]
$$

thus we obtain

$$
\begin{equation*}
\partial_{i} \frac{h}{\ell}(x)=\frac{\ell(x) \partial_{i} h(x)-h(x) \partial_{i} \ell(x)}{\ell^{2}(x)} \tag{A.19}
\end{equation*}
$$

Theorem A.3.21 (Composition). Let $A \subseteq \mathbf{R}^{m}$ and $B \subseteq \mathbf{R}^{n}$ be any sets and define maps $f: A \rightarrow B$ and $g: B \rightarrow R^{k}$. Let $f$ and $g$ be such that $f$ is differentiable at $x \in A$ and $g$ is differentiable at $f(x) \in B$. Then the map $g \circ f: A \rightarrow \mathbf{R}^{k}$ is differentiable at $x \in A$. Moreover, the differential $\mathrm{d}(g \circ f)(x): T \mathbf{R}_{x}^{m} \rightarrow T \mathbf{R}^{k} g(f(x))$ is

$$
\begin{equation*}
\mathrm{d}(g \circ f)(x)=\mathrm{d} g(f(x)) \circ \mathrm{d} f(x) \tag{A.20}
\end{equation*}
$$

Also, if $x \in \operatorname{Int} A$ and $f(x) \in \operatorname{Int} B$, then the theorem can be rewritten in terms of the Jacobian matrix

$$
\left[\begin{array}{ccc}
\partial_{1}\left(g_{1} \circ f\right)(x) & \ldots & \partial_{m}\left(g_{1} \circ f\right)(x) \\
\vdots & \ddots & \vdots \\
\partial_{1}\left(g_{k} \circ f\right)(x) & \ldots & \partial_{m}\left(g_{k} \circ f\right)(x)
\end{array}\right]=\left[\begin{array}{ccc}
\partial_{1} g_{1}(f(x)) & \ldots & \partial_{n} g_{1}(f(x)) \\
\vdots & \ddots & \vdots \\
\partial_{1} g_{k}(f(x)) & \ldots & \partial_{n} g_{k}(f(x))
\end{array}\right]\left[\begin{array}{ccc}
\partial_{1} f_{1}(x) & \ldots & \partial_{m} f_{1}(x) \\
\vdots & \ddots & \vdots \\
\partial_{1} f_{n}(x) & \ldots & \partial_{m} f_{n}(x)
\end{array}\right]
$$

hence we find the relation - for $1 \leqslant i \leqslant m$ and $1 \leqslant j \leqslant k$

$$
\begin{equation*}
\partial_{i}\left(g_{j} \circ f\right)(x)=\sum_{1 \leqslant r \leqslant n}\left(\partial_{r} g_{j}(f(x))\right) \cdot\left(\partial_{i} f_{r}(x)\right) \tag{A.21}
\end{equation*}
$$

Proof. Consider the filter base given by $h \rightarrow 0$, where $f(x+h)-f(x)=\mathrm{d} f(x)(h)+o(h)$, and as $t \rightarrow 0$ we find $g(f(x)+t)-g(f(x))=\mathrm{d} g(f(x))+o(t)$. Therefore by defining $t:=f(x+h)-f(x)$ we can write - as $h, t \rightarrow 0$

$$
\begin{aligned}
g(f(x+h))-g(f(x)) & =g(f(x)+t)-g(f(x)) \\
& =\mathrm{d} g(f(x))(t)+o(t) \\
& =\mathrm{d} g(f(x))(f(x+h)-f(x))+o(f(x+h)-f(x)) \\
& =\mathrm{d} g(f(x))(\mathrm{d} f(x)(h)+o(h))+o(f(x+h)-f(x)) \\
& =\mathrm{d} g(f(x))(\mathrm{d} f(x)(h))+\mathrm{d} g(f(x))(o(h))+o(f(x+h)-f(x))
\end{aligned}
$$

From definition we have that - as $h \rightarrow 0$

$$
\mathrm{d} g(f(x))(o(h))=g(f(x)+o(h))-g(f(x))-o(h)
$$

thus $\mathrm{d} g(f(x))(o(h)) \rightarrow 0$ as $h \rightarrow 0$, that is, $\mathrm{d} g(f(x))(o(h))=o(h)$. Also, as $h \rightarrow 0$, we have - recalling Proposition A.3.5

$$
t=f(x+h)-f(x)=\mathrm{d} f(x)(h)+o(h)=O(h)+o(h)=O(h)
$$

Notice that $o(O(h))=o(h)$, thus $o(f(x+h)-f(x))=o(h)$, and hence we finally conclude that

$$
\begin{aligned}
g(f(x+h))-g(f(x)) & =\mathrm{d} g(f(x))(\mathrm{d} f(x)(h))+\mathrm{d} g(f(x))(o(h))+o(f(x+h)-f(x)) \\
& =\mathrm{d} g(f(x))(\mathrm{d} f(x)(h))+o(h)+o(h) \\
& =\mathrm{d} g(f(x))(\mathrm{d} f(x)(h))+o(h)
\end{aligned}
$$

which from definition implies that $g \circ f$ is differentiable at $x$ and $\mathrm{d}(g \circ f)(x)=$ $\mathrm{d} g(f(x))(\mathrm{d} f(x))$.

Definition A.3.22 (Directional derivative). Let $f: E \rightarrow \mathbf{R}$ and $x \in \operatorname{Int} E$, and $v \in T \mathbf{R}_{x}^{m}$. If the following limit exists

$$
\begin{equation*}
\mathrm{D}_{v} f(x):=\lim _{t \rightarrow 0} \frac{f(x+v t)-f(x)}{t} \tag{A.22}
\end{equation*}
$$

then the real quantity $\mathrm{D}_{v} f(x) \in T \mathbf{R}_{f(x)}$ is called the directional derivative of $f$ at the point $x$ evaluated at the vector $v$.

Definition A.3.23 (Gradient). Let $E \subseteq \mathbf{R}^{m}$ and $f: E \rightarrow \mathbf{R}$ be a real map. We define the gradient of $f$ as the vector field grad $f: \mathbf{R}^{m} \rightarrow \mathbf{R}^{m}$ written as

$$
\operatorname{grad} f=\left(\partial_{1} f, \ldots, \partial_{m} f\right)
$$

Corollary A.3.24. Let $E \subseteq \mathbf{R}^{m}$. If $f: E \rightarrow \mathbf{R}$ is differentiable at $x \in \operatorname{Int} E$, then

$$
\mathrm{D}_{v} f(x)=\mathrm{d} f(x)(v)=\sum_{1 \leqslant j \leqslant m} \partial_{j} f(x) v_{j}
$$

Written in terms of the euclidean inner product in $\mathbf{R}^{m}$ :

$$
\mathrm{D}_{v} f(x)=\langle\operatorname{grad} f, v\rangle
$$

Proof. Let $\ell: \mathbf{R} \rightarrow E$ be a map $\ell(t)=x+v t$. Then the composition $f \circ \ell$ is such that the differential at the point $t=0$ - that is $\mathrm{d}(f \circ \ell)(0): T \mathbf{R}_{0} \rightarrow T \mathbf{R}_{x}-$ is given by $\mathrm{d}(f \circ \ell)(0)=\mathrm{d} f(x) \circ \mathrm{d} \ell(0)$, from the composition theorem. Thus, if $x \in \operatorname{Int} E$, then

$$
\partial(f \circ \ell)(0)=\sum_{1 \leqslant r \leqslant m}\left(\partial_{r} f(x)\right) \cdot\left(\partial \ell_{r}(0)\right)
$$

Moreover, $\partial \ell_{r}(0)=\lim _{R \ni h \rightarrow 0} \frac{\ell_{r}(0+h)-\ell_{r}(0)}{h}=\lim _{h \rightarrow 0} \frac{\left(x_{r}+v_{r} h\right)-x_{r}}{h}=v_{r}$, which implies that

$$
\partial(f \circ \ell)(0)=\sum_{1 \leqslant r \leqslant m} \partial_{r} f(x) v_{r}=\mathrm{d} f(x)(v) .
$$

In order to make the connection to the directional derivative, it suffices to observe that since $f \circ \ell: \mathbf{R} \rightarrow \mathbf{R}$ then $\partial(f \circ \ell)$ is just the single variable derivative of the composite map - that is

$$
\partial(f \circ \ell)(0)=\lim _{\mathrm{R} \ni t \rightarrow 0} \frac{(f \circ \ell)(0+t)-f(\ell(0))}{t}=\lim _{\mathrm{R} \ni t \rightarrow 0} \frac{f(x+v t)-f(x)}{t}=\mathrm{D}_{v} f(x) .
$$

■
Let $f: E \rightarrow \mathbf{R}$ be a differentiable map at $x \in E$. Define $\phi$ to be the angle between $\operatorname{grad} f(x)$ and a given unit vector $u \in \mathbf{T R}_{x}^{m}$. Lets analyse the implication of certain choices of $u$ :

- If $u=\frac{\operatorname{grad} f(x)}{\|\operatorname{grad} f(x)\|}$ then $\phi=0$ and hence

$$
\mathrm{D}_{u} f(x)=\langle\operatorname{grad} f(x), u\rangle=\|\operatorname{grad} f(x)\|\|u\| \cos (\phi)=\|\operatorname{grad} f(x)\|,
$$

which indicates that for this choice of $u$ the value $\mathrm{D} f(x) \in \mathbf{R}$ is maximized.

- If $u$ is any vector perpendicular to $\operatorname{grad} f(x)$ then $\mathrm{D}_{u} f(x)=0$.
- If $u=-\frac{\operatorname{grad} f(x)}{\|\operatorname{grad} f(x)\|}$, then $\phi=\frac{\pi}{2}$ and thus

$$
\mathrm{D}_{u} f(x)=-\|\operatorname{grad} f(x)\|
$$

which is the minimum value for the directional derivative over a unit vector.

Theorem A.3.25 (Inverse mapping differential). Let $x, y \in \mathbf{R}^{n}$ be any points and consider the map $f: U_{x} \rightarrow V_{y}$, where $U_{x} \subseteq \mathbf{R}^{n}$ is a neighbourhood of $x$ and $V_{y} \subseteq \mathbf{R}^{n}$ is a neighbourhood of $y$ - and define $f(x)=y$. Let $f$ be continuous at $x$ and have a continuous inverse mapping $f^{-1}: V_{y} \rightarrow U_{x}$ at the point $y$. If the map $f$ is differentiable at $x$ and $\mathrm{d} f(x): T \mathbf{R}_{x}^{n} \rightarrow T \mathbf{R}_{y}^{n}$ has an inverse $\mathrm{d} f(x)^{-1}: T \mathbf{R}_{y}^{n} \rightarrow T \mathbf{R}_{x}^{n}$, then the map $f^{-1}$ is differentiable at $y$ and the differential $\mathrm{d} f^{-1}(y): T \mathbf{R}_{y}^{n} \rightarrow T \mathbf{R}_{x}^{n}$ is such that

$$
\mathrm{d} f^{-1}(y)=(\mathrm{d} f(x))^{-1}
$$

Proof. From the continuity of $f$, for any neighbourhood $V \subseteq V_{y}$ of $y$, the preimage $f^{-1}(V) \subseteq U_{x}$ is open. Let $y+t \in V$ be some element, then there exists $x+h \in U_{x}$ for which $f(x+h)=y+t$. Moreover, since $f$ is continuous at $x$, - as $h \rightarrow 0$ - we have $t=f(x+h)-y \rightarrow 0$ and since $f^{-1}$ is continuous at $y$, - as $t \rightarrow 0-$ we have $h=f^{-1}(y+t)-x$.

From the differentiability of $f$ at $x$ we have that - as $h \rightarrow 0$, from Proposition A.3.5

$$
\begin{equation*}
t=\mathrm{d} f(x)(h)+o(h)=O(h)+o(h)=O(h) . \tag{A.23}
\end{equation*}
$$

Since $\mathrm{d} f(x)$ is invertible, from Eq. (A.23) we find - as $h \rightarrow 0$

$$
\begin{equation*}
\left.(\mathrm{d} f(x))^{-1}(t)=(\mathrm{d} f(x))^{-1} \circ \mathrm{~d} f(x)\right)(h)+(\mathrm{d} f(x))^{-1}(o(h))=h+o(h) \tag{A.24}
\end{equation*}
$$

Where $(\mathrm{d} f(x))^{-1}(o(h))=o(h)$ comes from Proposition A.3.3. Now let $\delta>0$ be such that, if $\|h\|<\delta$, then $\|o(h)\| \leqslant \frac{1}{2}\|h\|$. For such $\delta$ we find that - since $(\mathrm{d} f(x))^{-1}(t) \geqslant$ $h-o(h)$

$$
\left\|(\mathrm{d} f(x))^{-1}(t)\right\|=\|h-o(h)\| \geqslant\|h\|-\|o(h)\| \geqslant \frac{1}{2}\|h\|
$$

Hence, using the fact that $h=f^{-1}(y+t)+x \rightarrow 0$ as $t \rightarrow 0$, then the inequality obtained above - that is

$$
\|h\| \leqslant 2\left\|(\mathrm{~d} f(x))^{-1}(t)\right\|
$$

together with the fact that $(\mathrm{d} f(x))^{-1}(t) \rightarrow 0$ as $t \rightarrow 0$ - since linear maps are continuous (see Proposition A.2.37) and vanish at 0 - we conclude that $h=O(t)$. Therefore $h$ and $t$ have the same order over $h, t \rightarrow 0$, that is $h \asymp t$ and, equivalently, there exists $a, b>0$ constants for which exists a neighbourhood $U \subseteq \mathbf{R}^{n}$ of 0 such that $a\|t\| \leqslant\|h\| \leqslant b\|t\|$ - see Definition A.1.4. This shows that $o(h)=o(t)$. Henceforth, by the use of Eq. (A.24) we find

$$
f^{-1}(y+t)+f^{-1}(y)=(\mathrm{d} f(x))^{-1}(t)+o(t)
$$

Thus indeed $f^{-1}$ is differentiable at $y$ and $\mathrm{d} f^{-1}(x)=(\mathrm{d} f(x))^{-1}$.

## Mean Value Theorem

Remark A.3.26. In what remains of this chapter we'll denote a general domain by $G \subseteq \mathbf{R}^{m}$. Moreover, if $x, y \in \mathbf{R}^{n}$, we'll denote by $[x, y]$ the line segment $\gamma([0,1])$ where $\gamma:[0,1] \rightarrow \mathbf{R}^{n}$ maps $t \stackrel{\gamma}{\mapsto}(t-1) x+t y$. On the other hand, $(x, y)$ denotes $\gamma((0,1))$.

Theorem A.3.27 (Mean Value Theorem for real valued maps). Let $f: E \rightarrow \mathbf{R}$ be a map. If $f$ is continuous on the line segment $[x, y]$ and differentiable on $(x, y)$ - where $x, y \in E$ - then there exists $z \in(x, y)$ such that

$$
f(y)-f(x)=\mathrm{d} f(z)(y-x)
$$

Proof. Define $h(t)=(1-t) x+t y$. Let $g:[0,1] \rightarrow \mathbf{R}$ defined by $g=f \circ h$. Since $h$ and $f$ are both continuous maps on, respectively, $[0,1]$ and $h([0,1])=[x, y]$ - together with the fact that the composition of continuous maps yield a continuous map - we find that $g$ is continuous on $[0,1]$. Since $h$ and $f$ are differentiable at, respectively $(0,1)$ and $h((0,1))=(x, y)$ - by the composition of differentiable maps - we get that $g$ is differentiable on $(0,1)$.

We can now apply the Mean Value Theorem on $g$ to ensure that there exists $\bar{t} \in[0,1]$ such that

$$
g(1)-g(0)=g^{\prime}(\bar{t})
$$

Since $g^{\prime}(\bar{t})=[\mathrm{d} f(h(\bar{t}))]\left(h^{\prime}(\bar{t})\right)=\mathrm{d} f((1-\bar{t}) x+\bar{t} y)(y-x)$, making $z:=(1-\bar{t}) x+\bar{t} y \in(x, y)$ and substituting the values for $g$ in terms of $f$ we prove the theorem.

Corollary A.3.28 (Constant real valued map). Let $f: G \rightarrow \mathbf{R}$ be a map - where $G \subseteq \mathbf{R}^{m}$ is a domain. If $f$ is differentiable at every point of $G$ and its differential vanishes at all points of $G$, then $f$ is constant on the domain $G$.
Proof. Since for all $x \in G$ the differential $\mathrm{d} f(x)$ vanishes for all $v \in T \mathbf{R}_{x}^{m}$, then $\operatorname{dim} \operatorname{ker}(\mathrm{d} f(x))=m$ and $\mathrm{d} f(x)$ is the null mapping. Since $G$ is open, the partial derivatives $\partial_{j} f(x)$ exist - where $1 \leqslant j \leqslant m$ - and $\partial_{j} f(x)=0$.

Let $x \in G$ be any point and consider $B_{x}(r) \subseteq G$ an open ball centred in $x$. Let $y \in B_{x}(r)$ be any point, then from Theorem A.3.27 - since $[x, y] \subseteq B_{x}(r)$ - we have the existence of $z \in[x, y]$ such that

$$
\begin{equation*}
f(y)-f(x)=\mathrm{d} f(z)(y-x)=0 \tag{A.25}
\end{equation*}
$$

that is, $f(y)=f(x)$ for all points of the open ball - the map is constant at any open ball contained in $G$ of every point of $G$.

Let $x, y \in G$ be any points. Since $G$ is path-connected, there exists a continuous map $\gamma:[0,1] \rightarrow G$ such that $\gamma(0)=x$ and $\gamma(1)=y$. Consider any open ball $B_{x}(r) \subseteq G$ centred in $x$. Let $\Delta=\left\{\delta \in[0,1]: \gamma([0, \delta]) \in B_{x}(r)\right\}$. From the continuity of $\gamma$ and the fact that $\gamma(0)=x \in B_{x}(r)$, the collection $\Delta$ is non-empty. If $\delta \in \Delta-$ from Eq. (A.25) we find that $(f \circ \gamma)([0, \delta]=f(x)$. Since $\Delta$ is limited, we can define $\varepsilon=\sup \Delta$. From the continuity of $\gamma$ we have that $(f \circ \gamma)(\varepsilon)=f(x)$. We now show that, in fact, $\varepsilon=1$. Suppose - for the sake of contradiction - that $\varepsilon<1$, then there would exist some open ball $B_{\gamma(\varepsilon)}(d)$ such that $f\left(B_{\gamma(\varepsilon)}(d)\right)=f(\gamma(\varepsilon))=f(x)$ and, for instance, $\varepsilon+d \in \Delta$ and $\varepsilon<\varepsilon+d$, which contradicts the hypothesis that $\varepsilon=\sup \Delta$. From this, we conclude that $\varepsilon=1$ and hence $f(\gamma([0,1]))=f(x)=f(y)$ which closes the proof.
Lemma A.3.29. Let $[a, b] \subseteq \mathbf{R}$ be a closed interval and $f:[a, b] \rightarrow \mathbf{R}^{n}$ be a map. If $f$ is continuous on $[a, b]$ and differentiable on $(a, b)$, then there exists $x \in[a, b]$ such that

$$
\|f(b)-f(a)\|_{\mathbf{R}^{n}} \leqslant(b-a)\|\mathrm{d} f(x)\|_{\mathbf{R}^{n}}
$$

Proof. Let $\phi:[a, b] \rightarrow \mathbf{R}$ be defined as $\phi(t)=\langle f(b)-f(a), f(t)\rangle_{\mathbf{R}^{n}}$ - thus $\phi$ is continuous on $[a, b]$ and differentiable on $(a, b)$. Applying the Mean Value Theorem we ensure the existence of some $x \in(a, b)$ such that

$$
\begin{equation*}
\phi(b)-\phi(a)=(b-a) \phi^{\prime}(x)=(b-a)\langle f(b)-f(a), \mathrm{d} f(x)\rangle_{\mathbf{R}^{n}} . \tag{A.26}
\end{equation*}
$$

Since

$$
\begin{align*}
\phi(b)-\phi(a) & =\langle f(b)-f(a), f(b)\rangle_{\mathbf{R}^{n}}-\langle f(b)-f(a), f(a)\rangle_{\mathbf{R}^{n}} \\
& =\langle f(b)-f(a), f(b)-f(a)\rangle_{\mathbf{R}^{n}} \\
& =\|f(b)-f(a)\|_{\mathbf{R}^{n}}^{2} \tag{A.27}
\end{align*}
$$

We can substitute Eq. (A.27) into Eq. (A.26) to find

$$
\|f(b)-f(a)\|_{\mathbf{R}^{n}}^{2}=(b-a)\langle f(b)-f(a), \mathrm{d} f(x)\rangle_{\mathbf{R}^{n}} .
$$

From Hőlder's inequalities (see Proposition A.1.8) we know that $\langle v, u\rangle_{\mathbf{R}^{n}} \leqslant\|v\|_{\mathbf{R}^{n}}\|u\|_{\mathbf{R}^{n}}$. Using such fact and assuming $f(a) \neq f(b)$, we find that

$$
\|f(b)-f(a)\|_{\mathbf{R}^{n}}^{2}=(b-a)\langle f(b)-f(a), \mathrm{d} f(x)\rangle_{\mathbf{R}^{n}} \leqslant(b-a)\|f(b)-f(a)\|_{\mathbf{R}^{n}}\|\mathrm{~d} f(x)\|_{\mathbf{R}^{n}}
$$

implies $\|f(b)-f(a)\|_{\mathbf{R}^{n}} \leqslant(b-a)\|\mathrm{d} f(x)\|_{\mathbf{R}^{n}}$.
Theorem A.3.30 (Mean Value Theorem for $\mathbf{R}^{n}$ valued maps). Let $f: E \rightarrow \mathbf{R}^{n}$ be a continuous map on the line segment $[x, y]$ and differentiable on $(x, y)$ - with $x, y \in E$. If for all $z \in(x, y)$ and all $v \in T \mathbf{R}_{z}^{n}$ we have $\|\mathrm{d} f(z)(v)\|_{\mathbf{R}^{n}} \leqslant M\|v\|_{\mathbf{R}^{n}}$ - for some fixed $M \in \mathbf{R}$ - then

$$
\|f(y)-f(x)\|_{\mathbf{R}^{n}} \leqslant M\|y-x\|_{\mathbf{R}^{n}}
$$

Proof. Let $h:[0,1] \rightarrow E$ be defined as $h(t)=(1-t) x+t y$, and $g:[0,1] \rightarrow \mathbf{R}^{n}$ defined by $g=f \circ h$. Notice that $g$ is the composition of continuous maps on [0,1] and differentiable maps on $(0,1)$, so $g$ inherits both properties. By applying Lemma A.3.29 we ensure the existence of some $\bar{t} \in[0,1]$ such that

$$
\begin{equation*}
\|g(1)-g(0)\| \leqslant\|\mathrm{d} g(\bar{t})\| \tag{A.28}
\end{equation*}
$$

From the composition theorem we have that - for any $t \in(0,1)$

$$
\mathrm{d} g(t)=[\mathrm{d} f(h(t))](\mathrm{d} h(t))=[\mathrm{d} f(h(t))](y-x)
$$

Since $\|\mathrm{d} f(z)(v)\| \leqslant M\|v\|$, we get

$$
\begin{equation*}
\|\mathrm{d} g(t)\|=\|\mathrm{d} f(h(t))(y-x)\| \leqslant M\|y-x\| . \tag{A.29}
\end{equation*}
$$

By substituting Eq. (A.29) into Eq. (A.28) and noting that $g(1)=f(y)$ and $g(0)=f(x)$ we can finally conclude that

$$
\|f(y)-f(x)\| \leqslant M\|y-x\|
$$

Corollary A.3.31. Let $f: G \rightarrow \mathbf{R}^{n}$ be a differentiable map in $G$. If $\mathrm{d} f(x)$ is the zero-map — that is, $\mathrm{d} f(x)(y)=0$ for all $y \in T \mathbf{R}_{x}^{m}$ - for all $x \in G$, then $f$ is constant.
Proof. Let $M=0$ and use Theorem A.3.30.

## Sufficient Condition for Differentiability

Theorem A.3.32. Let $x \in \mathbf{R}^{m}$ be any point and let $U \subseteq \mathbf{R}^{m}$ be a neighbourhood of $x$. Let $f: U \rightarrow \mathbf{R}$ be a map. If for all $1 \leqslant j \leqslant m$ the partial derivatives $\partial_{j} f$ exist for each point of $U$, and $\partial_{j} f$ is continuous at $x$, then $f$ is differentiable at $x$.

Proof. Since the collection of open balls in $\mathbf{R}^{m}$ forms a base for the standard real topology, we may assume that $U=B_{x}(r)$ for some $r>0$. Consider a point $x+h \in B_{x}(r)$, then for all $0 \leqslant k \leqslant m$ we have that $y_{k}=x+\sum_{j=1}^{k} h_{j} e_{j} \in B_{x}(r)$. Moreover, if $0 \leqslant \ell, k \leqslant m$ then the line segment $\left[y_{\ell}, y_{k}\right] \subseteq U$. Notice that

$$
\begin{align*}
f(x+h)-f(x) & =f(x+h)+\left[\sum_{k=1}^{m-1} f\left(x+\sum_{j=1}^{k} h_{i} e_{j}\right)-f\left(x+\sum_{j=1}^{k} h_{i} e_{j}\right)\right]-f(x) \\
& =\sum_{k=1}^{m}\left[f\left(x+\sum_{j=1}^{k} h_{j} e_{j}\right)-f\left(x+\sum_{j=1}^{k-1} h_{j} e_{j}\right)\right] \\
& =\sum_{k=1}^{m} f\left(y_{k}\right)-f\left(y_{k-1}\right) \tag{A.30}
\end{align*}
$$

Since $f$ has partial derivatives for every point of $U$, for each $1 \leqslant k \leqslant m$, we can apply the Mean Value Theorem A.3.27 for the pair of points points $y_{k}$ and $y_{k-1}$, to find a point $z_{k}=x+\theta_{k} h_{k} e_{k}$ belonging to the segment line $\left[y_{k-1}, y_{k}\right] \subseteq U-$ with $\theta_{j} \in[0,1]$ - such that

$$
\begin{equation*}
f\left(y_{k}\right)-f\left(y_{k-1}\right)=\partial_{k} f\left(z_{k}\right) h_{k} \tag{A.31}
\end{equation*}
$$

Substituting Eq. (A.31) into Eq. (A.30) we get

$$
\begin{equation*}
f(x+h)-f(x)=\sum_{k=1}^{m} \partial_{k} f\left(z_{k}\right) h_{k} \tag{A.32}
\end{equation*}
$$

Since all partial derivatives are continuous at $x$, for all $1 \leqslant k \leqslant m$ there exists a map $\alpha_{k}: U \rightarrow \mathbf{R}$ such that $\alpha_{k}\left(h_{k}\right) \rightarrow 0$ as $h_{k} \rightarrow 0$ and we can write

$$
\begin{equation*}
\partial_{k} f\left(z_{k}\right)=\partial_{k} f\left(x+\theta_{k} h_{k} e_{k}\right)=\partial_{k} f(x)+\alpha_{k}\left(h_{k}\right) \tag{A.33}
\end{equation*}
$$

Substituting Eq. (A.33) into Eq. (A.32) we get the following relation

$$
f(x+h)-f(x)=\sum_{k=1}^{m} \partial_{k} f(x) h_{k}+\alpha_{k}\left(h_{k}\right) h_{k}=\sum_{k=1}^{m} \partial_{k} f(x) h_{k}+\sum_{k=1}^{m} \alpha_{k}\left(h_{k}\right) h_{k}
$$

Notice that as $h \rightarrow 0$ we have that $\alpha_{k}\left(h_{k}\right) h_{k} \rightarrow 0$ for all $1 \leqslant k \leqslant m$, hence $\alpha_{k}\left(h_{k}\right) h_{k}=$ $o(h)$ and thus

$$
f(x+h)-f(x)=\sum_{k=1}^{m} \partial_{k} f(x) h_{k}+o(h)
$$

From Definition A.3.7 the map $f$ is differentiable at $x$ - and $\mathrm{d} f(x)(h)=\sum_{k=1}^{m} \partial_{k} f(x) h_{k}$.

Definition A.3.33 (Continuously differentiable map). We denote by $C^{1}(G, \mathbf{R})$ the vector space of real valued continuously differentiable maps $\mathbf{R}^{m} \supseteq G \rightarrow \mathbf{R}$ - that is, with continuous partial derivatives in $G$.

## Higher Order Partial Derivatives

Definition A.3.34 ( $k$-th partial derivative). Let $f: G \rightarrow \mathbf{R}$ be a map - over a domain $G \subseteq \mathbf{R}^{m}$ - partially differentiable over its $j_{1}$-th variable. If in turn $\partial_{j_{1}} f: G \rightarrow \mathbf{R}$ is partially differentiable over its $j_{2}$-th variable, we say that $\partial_{j_{2}}\left(\partial_{j_{1}} f\right)=\partial_{j_{2} j_{1}} f: G \rightarrow \mathbf{R}$ is said to be the second partial derivative of $f$ with respect to $\left(j_{1}, j_{2}\right)$ variables. If such property holds $k$ times over variables the $\left(j_{1}, j_{2}, \ldots, j_{k}\right)$, we say that

$$
\partial_{j_{k} j_{k-1} \ldots j_{2} j_{1}} f: G \rightarrow \mathbf{R}
$$

is the $k$-th partial derivative of $f$ with respect to variables $\left(j_{i}\right)_{1 \leqslant i \leqslant k}$.
Definition A.3.35 (k-continuously differentiable). A map $f: G \rightarrow \mathbf{R}$ is said to be $k$ continuously differentiable if for all tuples of indices $I_{i}=\left(j_{1}, \ldots, j_{i}\right)$ and for all $1 \leqslant i \leqslant k$ the map $\partial_{I_{i}} f$ exists and is continuous at $G$.

Lemma A.3.36. Let $E \subseteq \mathbf{R}^{m}$ be an open set and $f: E \rightarrow \mathbf{R}$ be a map with existing partial derivatives $\partial_{i} f$ and $\partial_{j i} f$ at every point of $E$. Let $y \in E$ and a map $\phi_{j i}: V \rightarrow \mathbf{R}$ - where $V=\left\{(t, h) \in \mathbf{R}^{2}: y+t e_{i}+h e_{j} \in E\right\}$ - being defined by

$$
\phi_{j i}(t, h)=f\left(y+t e_{i}+h e_{j}\right)-f\left(y+t e_{i}\right)+f\left(y+h e_{j}\right)-f(y)
$$

Then for all pairs $(t, h) \in V-t, h \neq 0-$ there exists a point $x=y+\theta_{i} t e_{i}+\theta_{j} h e_{j} \in E$, where $\theta_{i}, \theta_{j} \in(0,1)$, such that

$$
\phi(t, h)=t h\left[\partial_{21} f(x)\right] .
$$

Proof. Since $E$ is open, $V$ is non-empty. Choose any non-zero tuple $(t, h) \in V$. Let the map $\omega:[0,1] \rightarrow \mathbf{R}$ be defined by

$$
\omega(\theta)=f\left(y+\theta t e_{i}+h e_{j}\right)-f\left(y+\theta t e_{i}\right)
$$

Notice that $\omega$ is continuous on $[0,1]$ and differentiable on $(0,1)$ - since $f$ has partial derivative $\partial_{i} f$. We can use the Mean Value Theorem to find $\theta_{i} \in(0,1)$ such that

$$
\begin{align*}
\omega(1)-\omega(0) & =\omega^{\prime}\left(\theta_{i}\right) \\
& =\left[\frac{\partial}{\partial \theta}\left(f\left(y+\theta t e_{i}+h e_{j}\right)-f\left(y+\theta t e_{i}\right)\right)\right]_{\theta=x_{1}} \\
& =\left[\theta \partial_{i} f\left(y+\theta t e_{i}+h e_{j}\right)-\theta \partial_{i} f\left(y+\theta t e_{i}\right)\right]_{\theta=\theta_{1}}  \tag{A.34}\\
& =t\left(\partial_{1} f\left(y+\theta_{i} t e_{i}+h e_{j}\right)-\partial_{1} f\left(y+\theta_{i} t e_{i}\right)\right) \tag{A.35}
\end{align*}
$$

Since $\partial_{i} f$ is partially differentiable - with respect to its $j$-th variable - on every point of $E$, we conclude that $\partial_{1} f$ continuous on the line segment $\left[y, y+t e_{i}+h e_{j}\right] \subseteq E$ and
is differentiable on the line segment $\left(y, y+t e_{i}+h e_{j}\right) \subseteq E$. Let a map $\gamma:[0,1] \rightarrow \mathbf{R}$ defined by

$$
\gamma(\theta)=\partial_{1} f\left(y+\theta_{i} t e_{i}+\theta e_{j}\right)
$$

which, from $\partial_{i} f$ properties, is continuous on $[0,1]$ and differentiable on $(0,1)$. We can apply the Mean Value Theorem on $\gamma$ to find an element $\theta_{j} \in(0,1)$ for which

$$
\begin{align*}
\gamma(1)-\gamma(0) & =\gamma^{\prime}\left(\theta_{j}\right) \\
& =\left[\frac{\partial}{\partial \theta} \partial_{1} f\left(y+\theta_{i} t e_{i}+\theta e_{j}\right)\right]_{\theta=\theta_{j}} \\
& =\left[h \partial_{21} f\left(y+\theta_{i} t e_{i}+\theta h e_{j}\right)\right]_{\theta=\theta_{j}} \\
& =h \partial_{21} f\left(y+\theta_{i} t e_{i}+\theta_{j} h e_{j}\right) \tag{A.36}
\end{align*}
$$

Therefore, substituting Eq. (A.36) into Eq. (A.34) we get

$$
\phi(t, h)=\omega(1)-\omega(0)=\operatorname{th}\left[\partial_{21} f\left(y+\theta_{i} t e_{i}+\theta_{j} h e_{j}\right)\right] .
$$

Theorem A.3.37 (Order of the partial derivative variables). Let $f: G \rightarrow \mathbf{R}$ be a map with partial derivatives $\partial_{i j} f: G \rightarrow \mathbf{R}$ and $\partial_{j i} f: G \rightarrow \mathbf{R}$. If $\partial_{i j} f$ and $\partial_{j i} f$ are both continuous over the point $x \in G$, then

$$
\partial_{i j} f(x)=\partial_{j i} f(x) .
$$

Proof. Since $G$ is open, there exists $h \in \mathbf{R}^{m}$ such that $x+h \in G$ - where we impose that $h_{i}, h_{j} \neq 0$. Define the map

$$
\begin{equation*}
\Delta(f, h)=\left[f\left(x+h_{i} e_{i}+h_{j} e_{j}\right)-f\left(x+h_{i} e_{i}\right)\right]-\left[f\left(x+h_{j} e_{j}\right)-f(x)\right] \tag{A.37}
\end{equation*}
$$

Define maps $\phi_{j i}, \phi_{i j}:[0,1] \rightarrow \mathbf{R}$ by

$$
\begin{align*}
\phi_{j i}(\theta) & =f\left(x+\theta h_{i} e_{i}+h_{j} e_{j}\right)-f\left(x+\theta h_{i} e_{i}\right)  \tag{A.38}\\
\phi_{i j}(\theta) & =f\left(x+h_{i} e_{i}+\theta h_{j} e_{j}\right)-f\left(x+\theta h_{j} e_{j}\right) \tag{A.39}
\end{align*}
$$

Applying Lemma A.3.36 on Eq. (A.38) we find $\theta_{i}, \theta_{j} \in(0,1)$ such that

$$
\begin{equation*}
\Delta(f, h)=\phi_{j i}(1)-\phi_{j i}(0)=h_{j} h_{i}\left[\partial_{j i} f\left(x+\theta_{i} h_{i} e_{i}+\theta_{j} h_{j} e_{j}\right)\right] \tag{A.40}
\end{equation*}
$$

Analogously, we can apply Lemma A. 3.36 on Eq. (A.39) to find $\bar{\theta}_{i}, \bar{\theta}_{j} \in(0,1)$ such that

$$
\begin{equation*}
\Delta(f, h)=\phi_{i j}(1)-\phi_{i j}(0)=h_{i} h_{j}\left[\partial_{i j} f\left(x+\bar{\theta}_{i} h_{i} e_{i}+\bar{\theta}_{j} h_{j} e_{j}\right)\right] \tag{A.41}
\end{equation*}
$$

Equating both Eq. (A.40) and Eq. (A.41) we get

$$
h_{j} h_{i}\left[\partial_{j i} f\left(x+\theta_{i} h_{i} e_{i}+\theta_{j} h_{j} e_{j}\right)\right]=h_{i} h_{j}\left[\partial_{i j} f\left(x+\bar{\theta}_{i} h_{i} e_{i}+\bar{\theta}_{j} h_{j} e_{j}\right)\right]
$$

Since $h_{i}, h_{j} \neq 0$, we have

$$
\partial_{j i} f\left(x+\theta_{i} h_{i} e_{i}+\theta_{j} h_{j} e_{j}\right)=\partial_{i j} f\left(x+\bar{\theta}_{i} h_{i} e_{i}+\bar{\theta}_{j} h_{j} e_{j}\right) .
$$

Since $\partial_{j i} f$ and $\partial_{i j} f$ are continuous at $x$, as $h \rightarrow 0$ we get

$$
\partial_{j i} f(x)=\partial_{i j} f(x)
$$

Corollary A.3.38. If $f \in C^{k}(G, \mathbf{R})$, then the maps $\partial_{i_{k} \ldots i_{1}} f$ for any $k$-tuple $I_{k}=\left(i_{1}, \ldots, i_{k}\right)$ - where $1 \leqslant i_{j} \leqslant m$ for all $1 \leqslant j \leqslant k$ - are the same for any permutation of $I_{k}$.

Proof. We proceed by induction on $k$. If $k=2$ then the proposition is true, from Theorem A.3.37. Assume the proposition is true for some $k>2$. Let $\sigma$ be any permutation on $I_{k}$, since any permutation can be written as the composition of finitely many elementary transpositions, it suffices to observe that - from the inductive hypothesis

$$
\begin{equation*}
\partial_{I_{k+1}} f=\partial_{i_{k+1}}\left(\partial_{I_{k}} f\right)=\partial_{i_{k+1}}\left(\partial_{\sigma\left(I_{k}\right)} f\right) . \tag{A.42}
\end{equation*}
$$

From Theorem A.3.37 following relation is verified

$$
\begin{equation*}
\partial_{I_{k+1}} f=\partial_{i_{k+1} i_{k}}\left(\partial_{I_{k-1}} f\right)=\partial_{i_{k} i_{k+1}}\left(\partial_{I_{k-1}} f\right)=\partial_{i_{k} i_{k+1} i_{k-1} \ldots i_{1}} f \tag{A.43}
\end{equation*}
$$

Thus, with Eq. (A.42) and Eq. (A.43) we find that the the proposition is true for $k+1$. This finishes the induction proof.

Example A.3.39. Let $f \in C^{k}(G, \mathbf{R})$ and, for some given $x \in G$, let $h \in \mathbf{R}^{m}$ be such that $x+h \in G$. Consider the map $\phi:[0,1] \rightarrow \mathbf{R}$, defined as

$$
\phi(t)=f(x+t h)
$$

Then $\phi \in C^{k}([0,1], \mathbf{R})$ and its $k$-th derivative is given by

$$
\phi^{(k)}(t)=\left(h_{1} \partial_{1}+\cdots+h_{m} \partial_{m}\right)^{k} f(x+t h) .
$$

More generally, for all $1 \leqslant j \leqslant k$ we have

$$
\phi^{(j)}(t)=\left(h_{1} \partial_{1}+\cdots+h_{m} \partial_{m}\right)^{j} f(x+t h)
$$

Definition A.3.40 (Hessian matrix). Let $f: E \rightarrow \mathbf{R}$ be twice partially differentiable with respect to all variables at the point $x \in \operatorname{Int} E$. The Hessian of $f$ is defined to be the matrix

$$
\text { Hess } f(x)=\left[\begin{array}{ccc}
\partial_{11} f(x) & \ldots & \partial_{1 m} f(x) \\
\vdots & \ddots & \vdots \\
\partial_{m 1} f(x) & \ldots & \partial_{m m} f(x)
\end{array}\right]
$$

Definition A.3.41. The Hessian of $f$ is defined to be the second order differential of $f$ at the interior point $x$ that is, the multilinear map

$$
\mathrm{d}^{2} f(x): T \mathbf{R}_{x}^{m} \times T \mathbf{R}_{x}^{m} \rightarrow \mathbf{R}, \text { mapping }(y, z) \stackrel{f}{\longmapsto} \sum_{i, j=1}^{m} \partial_{i j} f(x) y_{i} z_{i}
$$

## Taylor's Formula

Theorem A.3.42 (Taylor's formula). Let $x \in \mathbf{R}^{m}$ be a point and let $U \subseteq \mathbf{R}^{m}$ be a neighbourhood of $x$. Consider a point $h \in \mathbf{R}^{m}$ such that the line segment $[x, x+h]$ is contained in $U$. Let $f: U \rightarrow \mathbf{R}$ be a map $f \in C^{k+1}(U, \mathbf{R})$. Then the following equality holds

$$
f(x+h)-f(x)=\sum_{\ell=1}^{k} \frac{1}{\ell!}\left(h_{1} \partial_{1}+\cdots+h_{m} \partial_{m}\right)^{\ell} f(x)+r_{k}(x, h)
$$

Where the polynomial term is called Taylor polynomial of order $k$ of $f$ on $x$, and $r_{k}$ is the $k$-th order remainder of $f$ on $x$ - which can be written in the following forms:
(i) Integral form

$$
r_{k}(x, h)=\int_{0}^{1} \frac{(1-t)^{k}}{k!}\left(h_{1} \partial_{1}+\cdots+h_{m} \partial_{m}\right)^{k+1} f(x+t h) \mathrm{d} t
$$

(ii) Lagrange form - for some $\theta \in(0,1)$

$$
r_{k}(x, h)=\frac{1}{(k+1)!}\left(h_{1} \partial_{1}+\cdots+h_{m} \partial_{m}\right)^{k+1} f(x+\theta h)
$$

(iii) Peano form - as $h \rightarrow 0$

$$
r_{k}(x, h)=\frac{1}{(k+1)!}\left(h_{1} \partial_{1}+\cdots+h_{m} \partial_{m}\right)^{k+1} f(x)+o\left(\|h\|_{\mathbf{R}^{m}}^{k+1}\right) .
$$

and we rewrite the Taylor formula as

$$
f(x+h)-f(x)=\sum_{\ell=1}^{k+1} \frac{1}{\ell!}\left(h_{1} \partial_{1}+\cdots+h_{m} \partial_{m}\right)^{\ell} f(x)+o\left(\|h\|_{\mathbf{R}^{m}}^{k+1}\right) .
$$

Proof. Let $\phi:[0,1] \rightarrow \mathbf{R}$ be defined by $\phi(t)=f(x+t h)$. Notice that this implies in $\phi \in$ $C^{k+1}([0,1], \mathbf{R})$ and - using the Taylor's formula for one variable and Example A.3.39 - we find, for $\tau \in[0,1]$

$$
\phi(\tau)=f(x+\tau h)=\sum_{\ell=1}^{k} \frac{\phi^{(\ell)}(0)}{\ell!} \tau^{\ell}+R_{k}(\tau)
$$

Notice that $\phi(1)-\phi(0)=f(x+h)-f(x)$, hence

$$
f(x+h)-f(x)=\sum_{\ell=1}^{k} \frac{1}{\ell!}\left(h_{1} \partial_{1}+\cdots+h_{m} \partial_{m}\right)^{\ell} f(x)+r_{k}(x, h) .
$$

We can now analyse the possible forms for the remainders.
(i) (Integral form) We have $R_{k}$ given by

$$
R_{k}(\tau)=\int_{0}^{1} \frac{(1-t)^{k}}{k!} \phi^{(k+1)}(t \tau) t^{k+1} \mathrm{~d} t
$$

Hence

$$
r_{k}(x, h)=\int_{0}^{1} \frac{(1-t)^{k}}{k!}\left(h_{1} \partial_{1}+\cdots+h_{m} \partial_{m}\right)^{k+1} f(x+t \tau h) \tau^{k+1} \mathrm{~d} t
$$

(ii) (Lagrange form) For some $\theta \in(0,1), R_{k}$ is given by

$$
R_{k}(\tau)=\frac{1}{(k+1)!} \phi^{(k+1)}(\theta)
$$

Hence

$$
r_{k}(x, h)=\frac{1}{(k+1)!}\left(h_{1} \partial_{1}+\cdots+h_{m} \partial_{m}\right)^{k+1} f(x+\theta h)
$$

(iii) (Peano form) From the Lagrange form and as $h \rightarrow 0$

$$
\begin{aligned}
r_{k}(x, h) & =\frac{1}{(k+1)!}\left(h_{1} \partial_{1}+\cdots+h_{m} \partial_{m}\right)^{k+1} f(x+\theta h) \\
& =\frac{1}{(k+1)!}\left(h_{1} \partial_{1}+\cdots+h_{m} \partial_{m}\right)^{k+1} f(x)+o\left(\|h\|_{\mathbf{R}^{m}}^{k+1}\right)
\end{aligned}
$$

This finishes the proof.

## A. 4 Extrema on Several Variables

Definition A.4.1 (Extrema points). Let $f: E \rightarrow \mathbf{R}$ be a map. The map $f$ is said to have a local maximum (or local minimum) at $x_{0} \in \operatorname{Int} E$ if there exists a neighbourhood $U \subseteq E$ of $x_{0}$ for which $f(x) \leqslant f\left(x_{0}\right)$ (or $f(x) \geqslant f\left(x_{0}\right)$ ) for all $x \in U$. The local maximum (or minimum) is said to be strict if the strict inequality holds for all $x \in U \backslash x_{0}$. A local maximum (or minimum) is said to be a local extrema of $f$.

Theorem A.4.2 (Necessary condition for local extrema). Let $f: U \rightarrow \mathbf{R}$ be a map defined on the neighbourhood $U \subseteq \mathbf{R}^{m}$ of a point $x_{0}$. Assume that $f$ is partially differentiable at $x_{0}$. For $x_{0}$ to be a local extrema of $f$ it is necessary that for all $1 \leqslant j \leqslant m$ we have $\partial_{j} f\left(x_{0}\right)=0$, that is

$$
\operatorname{grad} f\left(x_{0}\right)=0
$$

Proof. Let $1 \leqslant j \leqslant m$ be any index and consider the map $g_{j}: \mathbf{R} \rightarrow \mathbf{R}$ given by

$$
x \stackrel{g_{j}}{\mapsto} f\left(x_{0}^{1}, \ldots, x_{0}^{j-1}, x, x_{0}^{j}, \ldots, x_{0}^{m}\right) .
$$

A necessary condition for $x_{0}^{j}$ to be an extrema of $g_{j}$ is that its derivative at $x_{0}^{j}$ needs to be zero. Moreover $g_{j}^{\prime}=\partial_{j} f$, thus the necessary condition for $g_{j}$ implies $\partial_{j} f\left(x_{0}\right)=0$. $\quad$,

Definition A.4.3. Let $A$ be a $(n \times n)$-symmetric matrix. We define the following:

- $A$ is said to be positive definite if for all $v \in \mathbf{R}^{n} \backslash 0$

$$
\langle v, A v\rangle>0 .
$$

On the other hand, $A$ is negative definite if for all $v \in \mathbf{R}^{n} \backslash 0$

$$
\langle v, A v\rangle<0 .
$$

- $A$ is positive semidefinite if for all $v \in \mathbf{R}^{n}$

$$
\langle v, A v\rangle \geqslant 0 .
$$

On the other hand, $A$ is negative semidefinite if for all $v \in \mathbf{R}^{n}$ we have

$$
\langle v, A v\rangle \leqslant 0 .
$$

Lemma A.4.4. Let $A$ be a symmetric $n \times n$ matrix. $A$ is positive definite if and only if there exists $\lambda>0$ such that for all $v \in \mathbf{R}^{n} \backslash 0$

$$
\langle v, A v\rangle \geqslant \lambda\|v\|^{2} .
$$

On the other hand, $A$ is negative definite if and only if there exists $\lambda>0$ such that for all $v \in \mathbf{R}^{n} \backslash 0$

$$
\langle v, A v\rangle \leqslant-\lambda\|v\|^{2} .
$$

Proof. Lets treat only the case for positive definite, the other is analogous and would be boring to repeat myself. Suppose there exists such $\lambda>0$ for which all $v \in \mathbf{R}^{n}$ satisfy $\langle v, A v\rangle \geqslant \lambda\|v\|^{2}$. Then since $v \neq 0$ and hence $\|v\|>0$, it follows immediately that $A$ is positive definite.

Suppose $A$ is positive definite. Consider the unitary $n-1$-sphere $S^{n-1}=\left\{v \in \mathbf{R}^{n}\right.$ : $\|v\|=1\}$. Since $S^{n-1}$ is closed and bounded, we conclude that $S^{n-1}$ is compact by Theorem A.2.9. Consider the map $f: S^{n-1} \rightarrow \mathbf{R}$ given by the mapping $v \mapsto\langle v, A v\rangle$ which is continuous. Therefore, from the global properties enunciated at Proposition A.2.36 we find that there exists $v_{0} \in S^{n-1}$ where $f$ assumes a minimum value. Let $f\left(v_{0}\right)=\lambda$ and since $A$ is positive definite, we have that $\lambda$ is necessarily positive. Let $v \in \mathbf{R}^{n} \backslash 0$ be any element, then from the definition of a minimum, we find $f\left(\frac{v}{\|v\|}\right) \geqslant \lambda$. Therefore we conclude that $\langle v, A v\rangle \geqslant \lambda\|v\|^{2}$, which finishes the proof.

Definition A.4.5 (Critical point). Let $f: U \rightarrow \mathbf{R}^{n}$ be a map defined on a neighbourhood $U \subseteq \mathbf{R}^{m}$ of a point $x_{0}$. Assume that $f$ is differentiable at $x_{0}$. We say that $x_{0}$ is a critical point of $f$ if the rank of the Jacobi matrix $f^{\prime}\left(x_{0}\right)$ has a rank less than $\min (m, n)$ - where $\min (m, n)$ is the maximum possible value of the rank.

Theorem A.4.6 (Classification of critical points of real valued maps). Let $f: U \rightarrow \mathbf{R}$ be twice continuously differentiable and let $x_{0} \in U \subseteq \mathbf{R}^{m}$ be an internal point such that $\operatorname{grad} f\left(x_{0}\right)=0$. If the matrix Hess $f\left(x_{0}\right)$ :

1. is positive definite, then $x_{0}$ is a local minimum of $f$.
2. is negative definite, then $x_{0}$ is a local maximum of $f$.
3. is indefinite, then $x_{0}$ is not an extremum point of $f$.

Proof. Let $h \in T_{x_{0}} \mathbf{R}^{m}$ and consider the Taylor polynomial of $f$ of order 2 at the point $x_{0}$ in the Peano form:

$$
f\left(x_{0}+h\right)-f\left(x_{0}\right)=\sum_{\ell=1}^{2} \frac{1}{\ell!}\left(h_{1} \partial_{1}+\cdots+h_{m} \partial_{m}\right)^{\ell} f\left(x_{0}\right)+o\left(\|h\|^{2}\right) .
$$

Notice however that for $x_{0}$ to be an extremum point candidate of $f$ we have from Theorem A.4.2 that grad $f\left(x_{0}\right)=0$ thus the Taylor polynomial only has its second order factor. Moreover, we can rewrite the second order term as an inner product of Hess $f\left(x_{0}\right)$ and $h$ in the following manner

$$
\sum_{i, j=1}^{m} \partial_{i j} f\left(x_{0}\right) h_{i} h_{j}=\left\langle h, \operatorname{Hess}\left(f\left(x_{0}\right)\right) h\right\rangle
$$

Thus we can now rewrite the Taylor polynomial in a way that lets us analyse the hessian of the map

$$
f\left(x_{0}+h\right)-f\left(x_{0}\right)=\left\langle h, \operatorname{Hess}\left(f\left(x_{0}\right)\right) h\right\rangle+o\left(\|h\|^{2}\right) .
$$

Suppose Hess $f\left(x_{0}\right)$ is positive definite, then using Lemma A.4.4 we find some $\lambda>0$ such that for all $v \in \mathbf{R}^{n}$ we have $\left\langle v, \operatorname{Hess}\left(f\left(x_{0}\right)\right) v\right\rangle \geqslant \lambda\|v\|^{2}$. In particular, we can choose $\delta>0$ for which $\left|o\left(\|h\|^{2}\right)\right| \leqslant \frac{\lambda}{4}\|h\|^{2}$ for all $h \in T_{x_{0}} \mathbf{R}^{n}$ such that $\|h\| \leqslant \delta$. Then for all $\|h\| \leqslant \delta$ we have

$$
f\left(x_{0}+h\right)-f\left(x_{0}\right)=\frac{1}{2}\left\langle h, \operatorname{Hess}\left(f\left(x_{0}\right)\right) h\right\rangle+o\left(\|h\|^{2}\right) \geqslant \frac{1}{2} \lambda\|h\|^{2}-\frac{\lambda}{4}\|h\|^{2} \geqslant \frac{\lambda}{4}\|h\|^{2}
$$

And since $\frac{\lambda}{4}\|h\|^{2}>0$ then $f\left(x_{0}+h\right)-f\left(x_{0}\right)>0$ and therefore - at least in the neighbourhood $B_{x_{0}}(\delta)$ - we are ensured that $x_{0}$ is a local minimum of $f$.

If on the other hand we have Hess $f\left(x_{0}\right)$ negative definite, then $\operatorname{Hess}\left(-f\left(x_{0}\right)\right)$ is positive definite and the proposition follows.

Suppose Hess $f\left(x_{0}\right)$ is indefinite and let $u \in S^{n-1}$ be the point where the mapping given by $\ell \mapsto\left\langle\ell, \operatorname{Hess}\left(f\left(x_{0}\right)\right) \ell\right\rangle$ assumes its minimum $m<0$ and $v \in S^{n-1}$ be the point where the map assumes it maximum $M>0$. Define $t>0$ so that $x_{0}+u t \in U$, then

$$
f\left(x_{0}+u t\right)-f\left(x_{0}\right)=\left\langle u t, \operatorname{Hess}\left(f\left(x_{0}\right)\right) u t\right\rangle+o\left(\|u t\|^{2}\right)=\frac{1}{2!} m t^{2}+o\left(t^{2}\right)
$$

Then, for $t \rightarrow 0$ we'll have $f\left(x_{0}+t u\right)-f\left(x_{0}\right)<0$ in some neighbourhood of $x_{0}$, since $m<0$ - this implies that $x_{0}$ is a local minimum of such neighbourhood. Moreover, if we now set $t>0$ so that $x_{0}+v t \in U$, it follows that

$$
f\left(x_{0}+v t\right)-f\left(x_{0}\right)=\frac{1}{2!}\left\langle v t, \operatorname{Hess}\left(f\left(x_{0}\right)\right) v t\right\rangle+o\left(\|v t\|^{2}\right)=\frac{1}{2!} M t^{2}+o\left(t^{2}\right)
$$

so that, as $t \rightarrow 0$ we have $f\left(x_{0}+t v\right)-f\left(x_{0}\right)>0$ for some neighbourhood of $x_{0}$, since $M>0$ - which now implies that $x_{0}$ is in fact a local maximum in some neighbourhood, which is a direct contradiction to the assertion that $x_{0}$ was a local minimum. This shows us that for Hess $f\left(x_{0}\right)$ indefinite, $x_{0}$ is not an extremum of $f$.

## A. 5 Implicit Map Theorem

Definition A.5.1 (Level curve). Let $f: \Omega \rightarrow \mathbf{R}$, where $\Omega \subseteq \mathbf{R}^{n}$, be any map. For every $c \in \mathbf{R}$ we define the set $N_{f}(c)=\{x \in \Omega: f(x)=c\}$ to be the $c$-level curve of $f$. Moreover, if $n=2$, we can call $N_{f}(c)$ the $c$-contour line of $f$ with value $c$.

Theorem A.5.2 (Implicit Theorem). Let $V$ and $L$ be normed vector spaces and $W$ be a Banach space. Define $\Omega \subseteq V \times W$ to be an open set and $\left(x_{0}, y_{0}\right) \in \Omega$. Consider $F: \Omega \rightarrow L$ to be a mapping such that

- $F\left(x_{0}, y_{0}\right)=c$ for some $c \in L$.
- $F$ is continuous at $\left(x_{0}, y_{0}\right)$.
- $F$ is differentiable and its differential $\mathrm{d} F: \Omega \rightarrow L$ is continuous at $\left(x_{0}, y_{0}\right)$.
- $\partial_{2} F\left(x_{0}, y_{0}\right): W \rightarrow L$ is an isomorphism, that is, it is invertible.

Then there exists neighbourhoods $U_{x_{0}} \subseteq V$ and $U_{y_{0}} \subseteq W$ such that $U_{x_{0}} \times U_{y_{0}} \subseteq \Omega$, and a map $f: U_{x_{0}} \rightarrow U_{y_{0}}$ such that

- $f\left(x_{0}\right)=y_{0}$.
- $f$ is continuous at $x_{0}$.
- $F(x, y)=0$ if and only if $f(x)=y$, for $x, y \in U_{x_{0}} \times U_{y_{0}}$.

Proof. To ease our lives, lets assume that $\Omega$ has the following form

$$
\Omega=\left\{(x, y) \in V \times W:\left\|x-x_{0}\right\|_{V}<\alpha \text { and }\left\|y-y_{0}\right\|_{W}<\beta\right\}
$$

If that is not the case, since $\Omega$ is open - and hence $\left(x_{0}, y_{0}\right)$ is an internal point - we can merely choose an open set contained in $\Omega$ that satisfies the above form.

Define a collection $\left\{g_{x}: x \in B_{x_{0}}(\alpha)\right\}$ of maps

$$
g_{x}(y)=y-\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}(F(x, y))
$$

The domain of each $g_{x}$ is defined to be the collection $B_{y_{0}}(\beta)=\left\{y \in W:\left\|y-y_{0}\right\|_{W}<\beta\right\}$. The maps $g_{x}$ are well defined since $\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}$ exists and is a continuous linear map - moreover, the domain of $g_{x}$ is the normed vector space $W$ - that is $g_{x}: B_{y_{0}}(\beta) \rightarrow W$.

Suppose $y_{x}$ is a fixed point of $g_{x}$, then $g_{x}\left(y_{x}\right)=y_{x}-\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}\left(F\left(x, y_{x}\right)\right)=y_{x}$ hence clearly $y_{x}$ is indeed a fixed point of $g_{x}$ if and only if $\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}\left(F\left(x, y_{x}\right)\right)=0$, that is, $F\left(x, y_{x}\right) \in \operatorname{ker}\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}$ - but $\partial_{2} F\left(x_{0}, y_{0}\right)$ is an isomorphism, so clearly $F\left(x, y_{x}\right)=0$.

Let $x \in B_{x_{0}}(\alpha)$ be any element and consider the map $g_{x}$. Notice that since $F$ is differentiable and $g_{x}$ is therefore a composition of differentiable maps, it follows that
$g_{x}$ is differentiable and since $\partial_{2} F\left(x_{0}, y_{0}\right)$ is continuous and linear we get

$$
\begin{aligned}
\partial_{2}\left(\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}(F(x, y))\right) & =\lim _{t \rightarrow 0} \frac{\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}(F(x, y+t))-\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}(F(x, y))}{t} \\
& =\lim _{t \rightarrow 0} \frac{\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}(F(x, y+t)-F(x, y))}{t} \\
& =\lim _{t \rightarrow 0}\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}\left(\frac{F(x, y+t)-F(x, y)}{t}\right) \\
& =\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}\left(\lim _{t \rightarrow 0} \frac{F(x, y+t)-F(x, y)}{t}\right) \\
& =\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}\left(\partial_{2} F(x, y)\right)
\end{aligned}
$$

Therefore we can write the differential of $g_{x}$ as

$$
\mathrm{d} g_{x}(y)=1_{W}-\partial_{2}\left(\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}(F(x, y))\right)=\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}\left(\partial_{2} F\left(x_{0}, y_{0}\right)-\partial_{2} F(x, y)\right)
$$

By the continuity of the map $\partial_{2} F\left(x_{0}, y_{0}\right)$ at the point $\left(x_{0}, y_{0}\right)$, we find that there exists $0<\gamma<\min (\alpha, \beta)$ such that in the neighbourhood $B_{x_{0}}(\gamma) \times B_{y_{0}}(\gamma) \subseteq \Omega$ we have

$$
\left\|\mathrm{d} g_{x}(y)\right\| \leqslant\left\|\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}\right\| \cdot\left\|\partial_{2} F\left(x_{0}, y_{0}\right)-\partial_{2} F(x, y)\right\|<\frac{1}{2} .
$$

Let $x \in B_{x_{0}}(\gamma)$ be any element and take any two $y_{1}, y_{2} \in B_{y_{0}}(\gamma)$. Then we have by the generalization of the mean value theorem that

$$
\left\|g_{x}\left(y_{1}\right)-g_{x}\left(y_{2}\right)\right\|_{W} \leqslant \sup _{t \in\left(y_{1}, y_{2}\right)}\|\mathrm{d} g(t)\|\left\|y_{1}-y_{2}\right\|_{W}<\frac{1}{2}\left\|y_{1}-y_{2}\right\|
$$

That is, $g_{x}$ is Lipschitz continuous.
From the definition of $g_{x}$ and the fact that $g_{x_{0}}\left(y_{0}\right)=y_{0}-\operatorname{since} F\left(x_{0}, y_{0}\right)=0$ - we have

$$
\begin{aligned}
\left\|g_{x}(y)-y_{0}\right\|_{W} & =\left\|g_{x}(y)-g_{x_{0}}\left(y_{0}\right)\right\|_{W} \\
& \leqslant\left\|g_{x}(y)-g_{x}\left(y_{0}\right)\right\|_{W}+\left\|g_{x}\left(y_{0}\right)-g_{x_{0}}\left(y_{0}\right)\right\|_{W} \\
& \leqslant \frac{1}{2}\left\|y-y_{0}\right\|_{W}+\left\|\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}\left(F\left(x, y_{0}\right)-F\left(x_{0}, y_{0}\right)\right)\right\|_{W} \\
& =\frac{1}{2}\left\|y-y_{0}\right\|_{W}+\left\|\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}\left(F\left(x, y_{0}\right)\right)\right\|_{W}
\end{aligned}
$$

Since $F$ is continuous at $\left(x_{0}, y_{0}\right)$ - so is the projection map $x \mapsto F\left(x, y_{0}\right)$ - then for all $\varepsilon^{\prime} \in(0, \gamma)$ there exists $\delta \in(0, \gamma)$ such that $\left\|x-x_{0}\right\|_{V}<\delta$ implies $\left\|F\left(x, y_{0}\right)-F\left(x_{0}, y_{0}\right)\right\|_{L}=$ $\left\|F\left(x, y_{0}\right)\right\|_{L}<\varepsilon^{\prime}$. In particular, choose $\varepsilon^{\prime}=\frac{\varepsilon}{2\left\|\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}\right\|}$ for any given $\varepsilon>0$, then for all $\left\|x-x_{0}\right\|_{V}<\delta$ and for all $\left\|y-y_{0}\right\|_{W} \leqslant \varepsilon$ we have - using the property that the
norm of linear maps is sub-multiplicative

$$
\begin{aligned}
\left\|g_{x}(y)-y_{0}\right\|_{W} & \leqslant \frac{1}{2}\left\|y-y_{0}\right\|_{W}+\left\|\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}\left(F\left(x, y_{0}\right)\right)\right\|_{W} \\
& \leqslant \frac{1}{2} \varepsilon+\left\|\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}\right\| \cdot\left\|F\left(x, y_{0}\right)\right\|_{W} \\
& \leqslant \frac{1}{2} \varepsilon+\left\|\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}\right\| \frac{\varepsilon}{2\left\|\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1}\right\|} \\
& =\frac{1}{2} \varepsilon+\frac{1}{2} \varepsilon=\varepsilon
\end{aligned}
$$

The relation above is equivalent to: for all $\left\|x-x_{0}\right\|_{V}<\delta$ we have

$$
g_{x}\left(\mathrm{Cl}\left(B_{y_{0}}(\varepsilon)\right)\right) \subseteq B_{y_{0}}(\varepsilon)
$$

Since $B_{y_{0}}(\varepsilon) \subseteq W$ is a closed set, it follows from Theorem A.1.17 that there exists a unique fixed point $y_{x} \in \operatorname{Cl}\left(B_{y_{0}}(\varepsilon)\right)$ of $g_{x}$. Define now the map $f: B_{x_{0}}(\delta) \rightarrow B_{y_{0}}(\varepsilon)$ where $f(x)=y_{x}$ and $y_{x}$ is the corresponding fixed point of each $g_{x}$.

From construction $B_{x_{0}}(\delta) \times B_{y_{0}}(\varepsilon) \subseteq \Omega$. Clearly, for all $x \in B_{x_{0}}(\delta)$ and $y \in B_{y_{0}}(\varepsilon)$, $F(x, y)=0$ if and only if $f(x)=y$. Moreover, we have immediately that $f\left(x_{0}\right)=y_{0}$. For the continuity of $f$ at the point $x_{0}$, we can observe that for all $\varepsilon \in(0, \gamma)$ there exists a $\delta \in(0, \gamma)$ such that $\left\|g_{x}\left(y_{x}\right)-y_{0}\right\|=\left\|f(x)-y_{0}\right\|<\varepsilon$, thus we are finally done.

We'll now extend the Implicit Map Theorem for cases 3 where we have more special conditions, allowing for additional properties for the implicit map $f$.

Lemma A.5.3 (Continuity of the implicit map). Let $F$ satisfy the properties described in Theorem A.5.2 and additionally suppose that there exists a neighbourhood of $\left(x_{0}, y_{0}\right)$ where the map $\partial_{2} F\left(x_{0}, y_{0}\right): W \rightarrow L$ is continuous. Then the implicit map $f: U \rightarrow V$ is such that there exists a neighbourhood of $x_{0}$ such that $f$ is continuous.

Proof.
Lemma A.5.4 (Differentiability of the implicit map). Let $F$ satisfy the properties described in Theorem A.5.2 and additionally suppose that the partial derivative $\partial_{1} F(x, y): V \rightarrow$ $L$ exists in some neighbourhood of the point $\left(x_{0}, y_{0}\right)$ and is continuous at $\left(x_{0}, y_{0}\right)$. Then the implicit map $f$ is differentiable at $x_{0}$ and

$$
\mathrm{d} f\left(x_{0}\right)=-\left[\partial_{2} F\left(x_{0}, y_{0}\right)\right]^{-1} \partial_{1} F\left(x_{0}, y_{0}\right)
$$

Proof.
Lemma A.5.5 (Continuous differentiability of the implicit map). Let $F$ satisfy the properties described in Theorem A.5.2 and additionally suppose that the partial derivatives of $F$ are continuous in some neighbourhood of $\left(x_{0}, y_{0}\right)$. Then the map $f$ is continuously differentiable in some neighbourhood of $x_{0}$ and its differential in this neighbourhood is given by

$$
\mathrm{d} f(x)=-\left[\partial_{1} F(x, f(x))\right]^{-1} \partial_{1} F(x, f(x)) .
$$

Lemma A.5.6 ( $C^{k}$ implicit map). Let $F$ satisfy the properties described in Theorem A.5.2 and additionally suppose that $F \in C^{k}(\Omega, L)$. Then the implicit map $f$ is a member of the class $C^{k}(U, W)$ in some neighbourhood $U \subseteq V$ of $x_{0}$.

Proof.
Write down the proof of the extensions of the basic Implicit Map Theorem

## Corollaries of the Implicit Map Theorem

## Inverse Map Theorem

Definition A.5.7 (Diffeomorphisms). Let $U$ and $V$ be open subsets of $\mathbf{R}^{m}$. A map $f: U \rightarrow V$ is said to be a isomorphism of manifolds of class $C^{p}$ (or diffeomorphism) for $p \in \mathbf{N} \cup\{\infty\}$ - if the following conditions are satisfied

- $f \in C^{p}(U, V)$.
- $f$ is bijective and $f^{-1} \in C^{p}(V, U)$.

Theorem A.5.8 (Inverse Map Theorem). Let $V$ and $W$ be Banach spaces and $\Omega \subseteq V$ be an open set. Consider a point $x_{0} \in \Omega$ and a map $f: \Omega \rightarrow W$ such that the following conditions are satisfied

- $f \in C^{1}(\Omega, W)$.
- $\mathrm{d} f\left(x_{0}\right)$ is invertible and $\left[\mathrm{d} f\left(x_{0}\right)\right]^{-1}$ is a continuous map.

Then there exists an open neighbourhood $X \subseteq \Omega$ of $x_{0}$ and an open neighbourhood $Y \subseteq W$ of the point $y_{0}=f\left(x_{0}\right)$, for which the restriction $f: X \rightarrow Y$ is bijective. The inverse $\operatorname{map} f^{-1}: Y \rightarrow X$ is differentiable and its differential is given by

$$
\mathrm{d} f^{-1}\left(y_{0}\right)=\left[\mathrm{d} f\left(x_{0}\right)\right]^{-1}
$$

Proof. Let $F: N \rightarrow W$ be a map defined on $N \subseteq V \times W$ where $N$ is a neighbourhood of $\left(x_{0}, y_{0}\right)$ and let $F(x, y)=f(x)-y$. Since $F$ is the composition of the restriction of maps that are continuously differentiable, it follows that $F$ is continuously differentiable, moreover, $\partial_{1} F\left(x_{0}, y_{0}\right)=\mathrm{d} f\left(x_{0}\right)$. Moreover, since $\mathrm{d} f\left(x_{0}\right)$ is invertible then $\partial_{1} F\left(x_{0}, y_{0}\right)$ is also invertible. We have $F\left(x_{0}, y_{0}\right)=0$ by construction, since $f\left(x_{0}\right)=y_{0}$.

We can now see that $F$ satisfies the requirements for the Implicit Map Theorem, thus there exists a neighbourhood $Y$ of $y_{0}$ and a continuously differentiable map $g: Y \rightarrow V$ (where we use extension Lemma A.5.6) for which $g(Y)$ is contained in a neighbourhood $X^{\prime} \subseteq V$ of $x_{0}$. Moreover, $F(x, y)=0$ if and only if $g(y)=x$, that is, $F(g(y), y)=0$ and therefore $f g(y)=y$ for any $y \in Y$, that is, $g$ is injective on $Y$ - also $g\left(y_{0}\right)=x_{0}$. The map $g$ has a differential given by (using the extension Lemma A.5.4)

$$
\mathrm{d} g(y)=\left[\partial_{1} F(x, y)\right]^{-1}\left[\partial_{2} F(x, y)\right], \text { for all }(x, y) \in X^{\prime} \times Y
$$

From the definition of $F$ we find that

$$
\mathrm{d} g(y)=[\mathrm{d} f(x)]^{-1}, \text { for all }(x, y) \in X^{\prime} \times Y
$$

Lets consider the restriction $f: g(Y) \rightarrow W$. Since $g$ is injective, the restriction $g: Y \rightarrow g(Y)$ is a bijection. Since $f$ is continuous and $Y$ is open, then $f^{-1}(Y)=g(Y)$ is open. Define $X=g(Y)$, so that $f: X \rightarrow Y$ is a bijection and clearly $g=f^{-1}$ for such restriction. Hence we conclude that

$$
\mathrm{d} f^{-1}(y)=[\mathrm{d} f(x)]^{-1}, \text { for all }(x, y) \in X \times Y
$$

Theorem A.5.9 (Open Map Theorem). Let $\Omega \subseteq \mathbf{R}^{n}$ be an open set and $f: \Omega \rightarrow \mathbf{R}^{n}$ be a continuously differentiable map. If $\mathrm{d} f(x)$ is invertible for all $x \in \Omega$, then the map $f$ is an open mapping - that is, maps open subsets of $\Omega$ to open subsets of $\mathbf{R}^{n}$.

Proof. Let $x \in \Omega$ be any point. From hypothesis, $\mathrm{d} f(x)$ is invertible, hence we can apply Theorem A.5.8 in order to obtain an open neighbourhood $V_{x} \subseteq \Omega$ of $x$ and $V_{f(x)} \subseteq \mathbf{R}^{n}$ such that the map $f: V_{x} \rightarrow V_{f(x)}$ is a local bijection and hence $f\left(V_{x}\right)$ is open. With this in our hands, we can create an open cover $\mathcal{U}=\left\{V_{x} \subseteq \Omega: x \in \Omega\right\}$ of such neighbourhoods - that is, given any open set $U \subseteq \Omega$, there exists a collection of open sets $\mathcal{V} \subseteq \mathcal{U}$ such that $U=\bigcup_{V \in \mathcal{V}} V$ and since $f(U)=\bigcup_{V \in \mathcal{V}} f(V)$ is the union of open sets, then $f(U)$ is open.

Theorem A.5.10 (Maximal Rank Theorem). Let $\Omega \subseteq \mathbf{R}^{n}$ be an open set and $x_{0} \in \Omega$. Let $f: \Omega \rightarrow \mathbf{R}^{m}$ be a continuously differentiable map. Define $y_{0} \in \mathbf{R}^{m}$ so that $f\left(x_{0}\right)=y_{0}$. The following holds
(a). Suppose $n \leqslant m$ and that $\mathrm{d} f\left(x_{0}\right)$ has maximal $\operatorname{rank}\left(\mathrm{d} f\left(x_{0}\right)\right)=n$. Then there exists open sets $\Omega_{y_{0}} \subseteq \mathbf{R}^{m}$ and $\Omega_{x_{0}} \subseteq \Omega \subseteq \mathbf{R}^{n}$, respectively neighbourhoods of the points $y_{0}$ and $x_{0}$ with $f\left(\Omega_{x_{0}}\right) \subseteq \Omega_{y_{0}}$, and a differentiable map $g: \Omega_{y_{0}} \rightarrow \mathbf{R}^{m}$ such that the following diagram commutes


Where $\iota: \mathbf{R}^{n} \hookrightarrow \mathbf{R}^{m}$ is the canonical inclusion map.
(b). Suppose $n \geqslant m$ and that $\mathrm{d} f\left(x_{0}\right)$ has maximal $\operatorname{rank}\left(\mathrm{d} f\left(x_{0}\right)\right)=m$. Then there exists $\Omega_{x_{0}} \subseteq \Omega \subseteq \mathbf{R}^{n}$, neighbourhood of $x_{0}$, and a differentiable map $g: \Omega_{x_{0}} \rightarrow \Omega$ such that the following diagram commutes


Where $\pi: \mathbf{R}^{n} \rightarrow \mathbf{R}^{m}$ is the canonical projection map.

Proof. 1. Since $\operatorname{rank}\left(\mathrm{d} f\left(x_{0}\right)\right)=n$, then, from the rank plus nullity theorem we find that $\operatorname{ker}\left(\mathrm{d} f\left(x_{0}\right)\right)=0$ and therefore $\mathrm{d} f\left(x_{0}\right)$ is injective. Consider the matrix representation $f^{\prime}\left(x_{0}\right)$ of the differential $\mathrm{d} f\left(x_{0}\right)$. From the injective property of $f$, there must exist $n$ linearly independent rows in $f^{\prime}\left(x_{0}\right)$. Let $A$ be the $(n \times n)$ matrix containing these linearly independent rows and $C$ the $((m-n) \times n)$-matrix containing the remaining rows of $f^{\prime}\left(x_{0}\right)$. Do row operations on $f^{\prime}\left(x_{0}\right)$ so that we see it as equivalent to the matrix

$$
\left[\begin{array}{l}
A \\
C
\end{array}\right]
$$

Notice that the collection of rows of $A$ form a basis for $\mathbf{R}^{n}$, thus $A$ is invertible and hence $\operatorname{det} A \neq 0$. Define a map $F: \Omega \times \mathbf{R}^{m-n} \rightarrow \mathbf{R}^{m}$ given by the mapping $(x, y) \mapsto f(x)+(0, y)$, then we obtain

$$
F^{\prime}\left(x_{0}, 0\right)=\left[\begin{array}{cc}
A & 0 \\
C & I_{m-n}
\end{array}\right]
$$

which in particular makes the, otherwise dependent, rows of $C$ into a collection of linearly independent vectors, by attaching the canonical base of the space $\mathbf{R}^{m-n}$ into each of them. This makes $F^{\prime}\left(x_{0}, 0\right)$ an invertible matrix. By applying Theorem A.5.8 we obtain a neighbourhood $U \subseteq \Omega \times \mathbf{R}^{m-n}$ of $\left(x_{0}, 0\right)$ and a neighbourhood $\Omega_{y_{0}} \subseteq \mathbf{R}^{m}$ of $F\left(x_{0}, 0\right)=f\left(x_{0}\right)=y_{0}$ for which the restriction map $F: U \rightarrow \Omega_{y_{0}}$ is an isomorphism of manifolds. Let $g: \Omega_{y_{0}} \rightarrow U$ be the continuously differentiable inverse of $F$, and define $\Omega_{x_{0}}=f^{-1}\left(\Omega_{y_{0}}\right) \cap \Omega$, which is clearly a neighbourhood of $x_{0}$. Notice that the composition $g f(x)=g F(x, 0)=(x, 0)=\iota(x)$, thus we are done.
2. Since $\operatorname{rank}\left(\mathrm{d} f\left(x_{0}\right)\right)$ equals the dimension of its codomain, it follows that $\mathrm{d} f\left(x_{0}\right)$ is a surjective linear map. Since $\mathrm{d} f\left(x_{0}\right)$ has rank $m$, then its matrix representation $f^{\prime}\left(x_{0}\right)$ has $m$ linearly independent columns. Let $D$ be the ( $m \times m$ )-matrix whose columns are those of $f^{\prime}\left(x_{0}\right)$ that are linearly independent and $C$ be the $(m \times(n-$ $m)$ )-matrix composed of the remaining columns of $f^{\prime}\left(x_{0}\right)$. Operate on the matrix $f^{\prime}\left(x_{0}\right)$ via column operations so that its final equivalent matrix is

$$
\left[\begin{array}{ll}
D & C
\end{array}\right]
$$

Since $D$ is composed of linearly independent vectors, $D$ is invertible. Define the map $F: \Omega \rightarrow \mathbf{R}^{m} \times \mathbf{R}^{n-m}$ by $(x, y) \mapsto(f(x), y)$. Clearly, $F$ is differentiable at $x_{0}$ and its matrix representation is

$$
F^{\prime}\left(x_{0}\right)=\left[\begin{array}{cc}
0 & I_{n-m} \\
D & C
\end{array}\right]
$$

which is invertible since the attachment of the canonical basis of $\mathbf{R}^{n-m}$ into the column vectors of $C$ transforms the collection of the last $n-m$ column vectors of $F^{\prime}\left(x_{0}\right)$ into a linearly independent set. Applying Theorem A.5.8 we are able to obtain a neighbourhood $\Omega_{x_{0}} \subseteq \Omega$ such that the restriction map $F: \Omega_{x_{0}} \rightarrow \mathbf{R}^{m} \times$
$\mathbf{R}^{n-m}$ is an isomorphism of manifolds. Let $g: \mathbf{R}^{m} \times \mathbf{R}^{n-m} \rightarrow \Omega_{x_{0}}$ be its continuously differentiable inverse map, then the composition $f g(x, y)=x=\pi(x, y)$ is merely the canonical projection map, as we expected.

Theorem A.5.11 (Constant rank). Let $U \subseteq \mathbf{R}^{n}$ be an open set, and $f: U \rightarrow \mathbf{R}^{m}$ be a $C^{\infty}$-map with a locally constant rank $k$ at a neighbourhood $V \subseteq U$ of a point $p \in U$. Then after a possible permutation of coordinates near $p$ and $f p$, the map $f$ assumes the form

$$
\left.f\right|_{U}=\left(f_{1}, \ldots, f_{k}, 0, \ldots, 0\right)
$$

In other words, there exists a $C^{\infty}$-isomorphism $\phi: Q \rightarrow \mathbf{R}^{n}$ centred at $p$ (that is, with $\phi p=0$ ), and a $C^{\infty}$-isomorphism $\psi: W \rightarrow \mathbf{R}^{m}$-for some neighbourhood $W \subseteq \mathbf{R}^{m}$ of $f p$-centred at $f p$ such that

$$
\begin{equation*}
\psi f \phi^{-1}\left(x_{1}, \ldots, x_{n}\right)=\left(x_{1}, \ldots, x_{k}, 0, \ldots, 0\right) \in \mathbf{R}^{m} \tag{A.44}
\end{equation*}
$$

Proof. Since $f$ has a locally constant rank $k$ at $p$, the local representation of the Jacobian matrix of $f$ at $V$ can be rearranged so that

$$
\operatorname{det}\left[\partial_{j} f_{i} x\right]_{1 \leqslant i, j \leqslant k} \neq 0
$$

for any $x \in V$. Define $\phi: U \rightarrow \mathbf{R}^{n}$ to be the map $\phi:=\left(f_{1}, \ldots, f_{k}, \pi_{k+1}, \ldots, \pi_{n}\right)$, so that one has a local Jacobian matrix at $V$ given by

$$
\operatorname{Jac} \phi=\left[\begin{array}{cccccc}
\partial_{1} f_{1} & \ldots & \partial_{k} f_{1} & \partial_{k+1} f_{1} & \ldots & \partial_{n} f_{1} \\
\vdots & \ddots & \vdots & \vdots & \ddots & \vdots \\
\partial_{1} f_{k} & \ldots & \partial_{k} f_{k} & \partial_{k+1} f_{k} & \ldots & \partial_{n} f_{k} \\
0 & \ldots & 0 & 1 & \ldots & 0 \\
\vdots & \ddots & \vdots & \vdots & \ddots & \vdots \\
0 & \ldots & 0 & 0 & \ldots & 1
\end{array}\right]
$$

Then $\operatorname{det}(\operatorname{Jac} \phi x)=\operatorname{det}\left[\partial_{j} f_{i} x\right]_{1 \leqslant i, j \leqslant k}$ is non-zero for every $x \in V$. Using the inverse map theorem, $\phi$ is a $C^{\infty}$-isomorphism $\phi: U_{p} \rightarrow U_{\phi p}$ for some neighbourhoods $U_{p} \subseteq V$ of $p$ and $U_{\phi p} \subseteq \mathbf{R}^{n}$ of $\phi p$, respectively.

Consider the map $g:=f \phi^{-1}: U_{\phi p} \rightarrow \mathbf{R}^{m}$. Since $\phi^{-1}$ has a constant maximal rank $n$ at $U_{\phi p}$, and $f: U_{p} \rightarrow \mathbf{R}^{m}$ has constant rank $k$ at $U_{p}$, it follows that $g$ has constant rank $k$ at $U_{\phi p}$. By construction, the Jacobian matrix of $g$ is

$$
\text { Jac } g=\left[\begin{array}{cccccc}
1 & \ldots & 0 & 0 & \ldots & 0 \\
\vdots & \ddots & \vdots & \vdots & \ddots & \vdots \\
0 & \ldots & 1 & 0 & \ldots & 0 \\
\partial_{1} g_{k+1} & \ldots & \partial_{k} g_{k+1} & \partial_{k+1} g_{k+1} & \ldots & \partial_{n} g_{k+1} \\
\vdots & \ddots & \vdots & \vdots & \ddots & \vdots \\
\partial_{1} g_{m} & \ldots & \partial_{k} g_{m} & \partial_{k+1} g_{m} & \ldots & \partial_{n} g_{m}
\end{array}\right]
$$

Therefore in the open set $U_{\phi p}$ one has $\partial_{j} g_{i}=0$ for every $k+1 \leqslant i \leqslant m$ and $k+1 \leqslant j \leqslant n$. Restricting the neighbourhood $U_{\phi p}$ to some inner open ball about $\phi p$-so that the neighbourhood is a convex set-we see that $g_{i}$ is independent of the last $m-k$ variables, for each $k+1 \leqslant i \leqslant m$.

Define a $C^{\infty}$-morphism $\psi: U_{\phi p} \rightarrow \mathbf{R}^{m}$ by making $\psi_{i}:=\pi_{i}$ for each $1 \leqslant i \leqslant k$, and $\psi_{i}:=\pi_{i}-g_{i}$ for each $k+1 \leqslant i \leqslant m$. The Jacobi matrix of $\psi$ has the form

$$
\operatorname{Jac} \psi=\left[\begin{array}{cccccc}
1 & \ldots & 0 & 0 & \ldots & 0 \\
\vdots & \ddots & \vdots & \vdots & \ddots & \vdots \\
0 & \ldots & 1 & 0 & \ldots & 0 \\
-\partial_{1} g_{k+1} & \ldots & -\partial_{k} g_{k+1} & 1 & \ldots & 0 \\
\vdots & \ddots & \vdots & \vdots & \ddots & \vdots \\
-\partial_{1} g_{m} & \ldots & -\partial_{k} g_{m} & 0 & \ldots & 1
\end{array}\right]
$$

Which has $\operatorname{det}(\operatorname{Jac} \psi)=1$-hence $\psi$ is a $C^{\infty}$-isomorphism $\psi: U_{f p} \rightarrow U_{\psi f p}$ for some neighbourhoods $U_{f p}, U_{\psi f p} \subseteq \mathbf{R}^{m}$ of $f p$ and $\psi f p$, respectively. Considering the neighbourhood $g^{-1} U_{f p}$ we find that the map

$$
\psi f \phi^{-1}: g^{-1} U_{f p} \longrightarrow \mathbf{R}^{m}
$$

is a $C^{\infty}$-morphism. Defining $Q:=\phi^{-1} U_{\phi p}$ and $W:=\psi^{-1} U_{\psi f p}$ we obtain induced $C^{\infty}{ }^{-}$ isomorphisms $\phi: Q \rightarrow \phi Q$ and $\psi: W \rightarrow \psi W$ such that $\psi f \phi^{-1}$ satisfies the required Eq. (A.44).

## A. 6 Extrema With Constraints

Theorem A.6.1 (Existence of Lagrange multipliers). Let $f: \Omega \rightarrow \mathbf{R}$ be a map, where $\Omega \subseteq \mathbf{R}^{d}$, and a map $F: \Omega \rightarrow \mathbf{R}^{m}$, where $m<d$ - which will be called constraint map. Suppose $x_{0}$ is a local extremum of $f$ in the surface $S \subseteq \Omega$ defined by the constraint $S=\{x \in \Omega: F(x)=0\}$. Suppose additionally that $\operatorname{rank} F^{\prime}(x)=m$ for every $x \in \Omega$. Then, there exists a vector $\lambda \in \mathbf{R}^{m}$ - whose components are called Lagrange multipliers - such that

$$
f^{\prime}\left(x_{0}\right)=F^{\prime}\left(x_{0}\right) \lambda
$$

In the particular case where $m=1$, the constraint map generates a 1 dimensional surface and therefore grad $f\left(x_{0}\right)=\lambda \operatorname{grad} F\left(x_{0}\right)$, where $\lambda \in \mathbf{R}$.

Proof. Since $F^{\prime}\left(x_{0}\right)$ has rank $m$, let $D$ be the $(m \times m)$-matrix whose columns are the $m$ linearly independent columns of $F^{\prime}\left(x_{0}\right)$. Define $C$ to be the $(m \times d-m)$-matrix whose columns are the remaining columns of $F^{\prime}\left(x_{0}\right)$. By means of column operations, arrange $F^{\prime}\left(x_{0}\right)$ into an equivalent matrix of the form

$$
\left[\begin{array}{ll}
C & D
\end{array}\right]
$$

That is, the equivalent matrix has an invertible principal minor $D$ defined on its last $m$ columns. If we now identify points $x \in \mathbf{R}^{d}$ with points $\left(x_{1}, x_{2}\right)=x \in \mathbf{R}^{d-m} \times \mathbf{R}^{m}$, we find that $\partial_{2} F\left(x_{0}^{1}, x_{0}^{2}\right)$ is an invertible map (it corresponds to the matrix $D$ ).

We can now apply the Implicit Map Theorem to find neighbourhoods $\Omega_{x_{0}^{1}} \subseteq \mathbf{R}^{d-m}$ and $\Omega_{x_{0}^{2}} \subseteq \mathbf{R}^{m}$ of $x_{0}^{1}$ and $x_{0}^{2}$, respectively, and a map $g: \Omega_{x_{0}^{1}} \rightarrow \Omega_{x_{0}^{2}}$ for which $F\left(x_{1}, x_{2}\right)=0$ if and only if $g\left(x_{1}\right)=x_{2}$. In particular, the mapping $x_{1} \mapsto F\left(x_{1}, g\left(x_{1}\right)\right)$ is identically zero and therefore $\Omega_{x_{0}^{1}} \times g\left(\Omega_{x_{0}^{1}}\right) \subseteq S$. From the fact that $x_{0}=\left(x_{0}^{1}, x_{0}^{2}\right)=\left(x_{0}^{1}, g\left(x_{0}^{1}\right)\right)$ is a local extremum of $f$, defining $\phi: \Omega_{x_{0}^{1}} \rightarrow \mathbf{R}$ by $\phi\left(x_{1}\right)=f\left(x_{1}, g\left(x_{1}\right)\right)$, we conclude that $x_{0}^{1}$ is a local maximum of $\phi$. In particular, it is necessary that $\mathrm{d} \phi\left(x_{0}^{1}\right)=0$ and therefore

$$
\mathrm{d} \phi\left(x_{0}^{1}\right)=\sum_{j=1}^{d-m} \partial_{j} f\left(x_{0}^{1}, g\left(x_{0}^{1}\right)\right)+\sum_{i=d-m+1}^{m} \partial_{i} f\left(x_{0}^{1}, g\left(x_{0}^{1}\right)\right) \partial_{i} g\left(x_{0}^{1}\right)=0 .
$$

If we now regard $f$ as map of the form $\mathbf{R}^{d-m} \times \mathbf{R}^{m} \rightarrow \mathbf{R}^{m}$, then we see that

$$
\mathrm{d} \phi\left(x_{0}^{1}\right)=\partial_{1} f\left(x_{0}^{1}, g\left(x_{0}^{1}\right)\right)+\partial_{2} f\left(x_{0}^{1}, g\left(x_{0}^{1}\right)\right) \mathrm{d} g\left(x_{0}^{1}\right)=0 .
$$

From the Implicit Map Theorem, $g$ was constructed so that

$$
\mathrm{d} g\left(x_{0}^{1}\right)=-\left[\partial_{2} F\left(x_{0}^{1}, g\left(x_{0}^{1}\right)\right)\right]^{-1} \partial_{1} F\left(x_{0}^{1}, g\left(x_{0}^{1}\right)\right) .
$$

Hence, substituting into the above equation we find

$$
\partial_{1} f\left(x_{0}^{1}, g\left(x_{0}^{1}\right)\right)-\partial_{2} f\left(x_{0}^{1}, g\left(x_{0}^{1}\right)\right)\left[\partial_{2} F\left(x_{0}^{1}, g\left(x_{0}^{1}\right)\right)\right]^{-1} \partial_{1} F\left(x_{0}^{1}, g\left(x_{0}^{1}\right)\right)=0
$$

Now, if we define $\lambda \in \mathbf{R}^{m}$ by $\lambda=\partial_{2} f\left(x_{0}^{1}, g\left(x_{0}^{1}\right)\right)\left[\partial_{2} F\left(x_{0}^{1}, g\left(x_{0}^{1}\right)\right)\right]^{-1}$, we find that

$$
\begin{aligned}
& \partial_{1} f\left(x_{0}^{1}, g\left(x_{0}^{1}\right)\right)=\partial_{1} F\left(x_{0}^{1}, g\left(x_{0}^{1}\right)\right) \lambda \\
& \partial_{2} f\left(x_{0}^{1}, g\left(x_{0}^{1}\right)\right)=\partial_{2} F\left(x_{0}^{1}, g\left(x_{0}^{1}\right)\right) \lambda
\end{aligned}
$$

So that we can conclude

$$
\mathrm{d} f\left(x_{0}\right)=\mathrm{d} F\left(x_{0}\right) \lambda
$$

Theorem A.6.2 (Sufficient condition for a constraint extremum). Let $\Omega \subseteq \mathbf{R}^{d}$ and maps $f: \Omega \rightarrow \mathbf{R}$, and $F: \Omega \rightarrow \mathbf{R}^{m}$, both of which are $C^{2}(\Omega, \mathbf{R})$. Let $S \subseteq \Omega$ be the surface defined by $S=\{x \in \Omega: F(x)=0\}$. Consider $x_{0} \in S$ as a possible candidate of local extremum of $f$ in the surface $S$. Suppose additionally that $\operatorname{rank} F^{\prime}(x)=m$ for every $x \in \Omega$. Define the Lagrange multipliers $\lambda \in \mathbf{R}^{m}$ so that the map $L: S \rightarrow \mathbf{R}$ given by

$$
L(x)=f(x)-\langle\lambda, F(x)\rangle
$$

satisfy $\operatorname{grad} L(x)=0$ for all $x \in S$ and $\partial_{\lambda} L(x)=0$.
It's sufficient for $x_{0} \in S$ to be a local extremum in $S$ if Hess $L\left(x_{0}\right)$ is either positive definite or negative definite. If Hess $L\left(x_{0}\right)$ is not definite, then $x_{0}$ cannot be an extremum point. Moreover, if Hess $L\left(x_{0}\right)$ is positive definite, then $x_{0}$ is a local minimum on $S$, on the other hand, if Hess $L\left(x_{0}\right)$ is negative definite, then $x_{0}$ is a local maximum on $S$.

Proof. Since $x_{0} \in S$, then in particular grad $L\left(x_{0}\right)=0$, thus, as $S \ni x \rightarrow x_{0}$ we have the polynomial approximation

$$
\begin{aligned}
L(x)-L\left(x_{0}\right) & =\frac{1}{2!} \sum_{i, j=1}^{d} \partial_{i j} L\left(x_{0}\right)\left(x_{i}-x_{0}^{i}\right)\left(x_{j}-x_{0}^{j}\right)+o\left(\left\|x-x_{0}\right\|^{2}\right) \\
& =\frac{1}{2!}\left\langle x-x_{0}, \operatorname{Hess}\left(L\left(x_{0}\right)\right)\left(x-x_{0}\right)\right\rangle+o\left(\left\|x-x_{0}\right\|_{\mathbf{R}^{d}}^{2}\right) .
\end{aligned}
$$

We'll assume that $S$ - which is a $(d-m)$-dimensional surface, since rank $F(x)=m$ - can be parametrically defined in some neighbourhood of $x_{0} \in S$ by a smooth map $\mathbf{R}^{d-m} \ni t \mapsto x(t) \in \mathbf{R}^{d}$ such that $x(0)=x_{0}$ and that exists a neighbourhood of $0 \in \mathbf{R}^{d-m}$ for which the parametrization is bijective. Since the mapping is smooth, as $t \rightarrow 0$ we have

$$
x(t)-x(0)=\mathrm{d} x(0)(t)+o\left(\|t\|_{\mathbf{R}^{d-m}}\right) .
$$

Which implies that as $t \rightarrow 0$ we have $\|x(t)-x(0)\|_{\mathbf{R}^{d}}=O\left(\|t\|_{\mathbf{R}^{d}}\right)$.
We can now exploit the parametrization of the surface $S$ so that, as $t \rightarrow 0$ we have

$$
\begin{aligned}
L(x(t))-L\left(x_{0}\right) & =\frac{1}{2!}\langle t, \operatorname{Hess}(L(x(0))) t\rangle+o\left(O\left(\|t\|_{\mathbf{R}^{d}}\right)\right) \\
& =\frac{1}{2!}\langle t, \operatorname{Hess}(L(x(0))) t\rangle+o\left(\|t\|_{\mathbf{R}^{d}}\right)
\end{aligned}
$$

Where the entries of the Hessian are of the form $\partial_{i j} L(x(0))=\partial_{i j} L(x(0)) \partial_{i} x(0) \partial_{j} x(0)$. Then, if Hess $L(x(0))$ is positive or negative definite, by Theorem A.4.6, we obtain that $t=0$ is an extremum of $L(x(t))$. On the other hand, since there exists a neighbourhood of $t=0$ for which $x(t)$ is a bijective parametrization, it follows that $L$ has an extremum at $x_{0}$ - and hence $x_{0}$ is an extremum of $f$ in $S$ and the classifications of maximum or minimum come again from the same theorem. If Hess $L(x(0))$ is indefinite, by Theorem A.4.6 we conclude that $L(x(t))$ has no extremum at $t=0-$ and with the same analogous arguments as before, we argue that $L$ has no extrema at $x_{0}$ and neither does $f$.

## A. 7 Riemann Integration of Real Valued Maps

## Primary Definitions

The main setting we are going to be working in this section, which will encompass the study of multiple Riemann integrals, is the standard euclidean space $\mathbf{R}^{n}$ and the $n$-dimensional closed intervals $I=[a, b]=\left\{x \in \mathbf{R}^{n}: a_{j} \leqslant x_{j} \leqslant b_{j}\right.$ for $\left.1 \leqslant j \leqslant n\right\}$. If it seems fit, we can denote that a point $x \in \mathbf{R}^{n}$ lies in the interval generated by given points $a, b \in \mathbf{R}^{n}$ by simply saying that $a \leqslant x \leqslant b$. Another terminology we are going to adopt is that, the interval $I$ is non-degenerate if $a_{j}<b_{j}$ for all $1 \leqslant j \leqslant n$.

Notation A.7.1. In this section we denote by $I^{n}$ the collection of all closed intervals of $\mathbf{R}^{n}$.

Definition A.7.2 (Interval measure). We define the map vol: $\mathcal{I}^{n} \rightarrow \mathbf{R}$ as the measure (or volume) of the $n$-dimensional closed intervals of $\mathbf{R}^{n}$, it's defined as the product of the interval sides, that is

$$
\operatorname{vol} I:=\prod_{j=1}^{n} b_{j}-a_{j}, \text { for } I=[a, b] \in \mathcal{I}^{n} .
$$

Corollary A.7.3 (Measure of intervals in $\mathbf{R}^{n}$ ). Let $I:=[a, b] \subseteq \mathbf{R}^{n}$ be an $n$-dimensional closed interval, then the following properties are satisfied concerning the measure vol:
(a) (Homogeneity) Let $\gamma \geqslant 0$ be a scalar and define the multiplication of the interval by $\gamma$ as $\gamma I:=[\gamma a, \gamma b]$. Then we have that

$$
\operatorname{vol}(\lambda I)=\lambda^{n} \operatorname{vol} I
$$

(b) (Additivity) Given a finite collection of closed intervals $\left\{I_{j} \subseteq \mathbf{R}^{n}\right\}_{j=1}^{p}$, we have that

$$
\operatorname{vol} \bigcup_{j=1}^{p} I_{j}=\sum_{j=1}^{p} \operatorname{vol} I_{j} .
$$

(c) (Cover inequality) Given a finite closed cover $\left\{I_{j}\right\}^{p}$, by $n$-dimensional closed intervals, of $I$ - that is $I \subseteq \bigcup_{j=1}^{p} I_{j}$ - then

$$
\operatorname{vol} I \leqslant \sum_{j=1}^{p} \operatorname{vol} I_{j}
$$

Definition A.7.4 (Partition). Let $I \subseteq \mathbf{R}^{n}$ be a closed interval. A partition on $I$ is a finite collection of closed intervals $\left\{I_{j}\right\}_{j=1}^{p}$ such that $I=\bigcup_{j=1}^{p} I_{j}$. The intervals pertaining to the partition are said to be finer than $I$.

Definition A.7.5 (Partition mesh). Given a partition $P \in 2^{I^{n}}$, we define the mesh of $P$ as the maximum diameter (recall Definition A.2.6) of the elements of the partition. That is, mesh: $2^{I^{n}} \rightarrow \mathbf{R}$ is a map defined by

$$
\operatorname{mesh}(P):=\max _{I \in P} d(I)
$$

Definition A.7.6 (Distinguished points). Given a partition $P=\left\{I_{j}\right\}_{j=1}^{p} \in 2^{I^{n}}$, we define a collection of distinguished points of the partition as a collection of points $\xi:=\left\{\xi_{j} \in\right.$ $\left.I_{j}\right\}_{j=1}^{p}$. The partition $P$ together with the distinguished points $\xi$ will be denoted as the pair $(P, \xi)$ - the collection of pairs $(P, \xi)$ will be denoted by $\mathcal{P}$.

An important filter base $\mathcal{B} \subseteq 2^{\mathcal{P}}$ is defined as the collection of sets $B_{d}$, where $d>0$ is a scalar, such that $B_{d}:=\{(P, \xi) \in \mathcal{P}: \operatorname{mesh}(P)<d\}$. We'll commonly denote $\mathcal{B}$ by $\operatorname{mesh}(P) \rightarrow 0$.

## Riemann Sums and Integrals

Definition A.7.7 (Riemann sum). Let $f: I \rightarrow \mathbf{R}$ be a map where $I \in I^{n}$. Consider the partition together with distinguished points $(P, \xi) \in \mathcal{P}$, then, we define the Riemann sum $\sigma$ : $\mathbf{R}^{I} \times \mathcal{P} \rightarrow \mathbf{R}$ by

$$
\sigma(f, P, \xi):=\sum_{j=1}^{p} f\left(\xi_{j}\right) \operatorname{vol}\left(I_{j}\right)
$$

where $P:=\left\{I_{j}\right\}_{j=1}^{p}$. We say that $\sigma(f, P, \xi)$ is the Riemann sum of the map $f$ with respect to the partition $P$ and distinguished points $\xi$.

Definition A.7.8 (Riemann integrable maps). A map $f: I \rightarrow \mathbf{R}$ is said to be Riemann integrable if the limit

$$
\lim _{\operatorname{mesh}(P) \rightarrow 0} \sigma(f, P, \xi)
$$

exists in R. We'll denote the $\mathbf{R}$-vector space of Riemann integrable maps with a given domain $E \subseteq \mathbf{R}^{n}$ by $\mathcal{R}(E)^{1}$.

Proposition A.7.9 (Boundness of Riemann integrable maps). Let $f: I \rightarrow \mathbf{R}$ be a Riemann integrable map. Then, $f$ is bounded on $I$.

Proof. We prove the contrapositive proposition. Suppose that $f$ is unbounded on $I$ and let $P$ be any partition of the interval $I$. In particular, since $P$ covers $I$, then there exists an interval $I_{k} \in P$ for which $f$ is unbounded. Let $\xi$ be any collection of distinguished points of $P$ and define $\xi^{\prime}$ as the collection of distinguished points $\xi_{j}^{\prime}:=\xi_{j}$ for $j \neq m$, and $\xi_{k}^{\prime} \in I_{k}$ to be such that $\xi_{k}^{\prime} \neq \xi_{k}$. Then, from construction, it follows that $\sigma(f, P, \xi)-\sigma\left(f, P, \xi^{\prime}\right)=\left(f\left(\xi_{k}\right)-f\left(\xi_{k}^{\prime}\right)\right) \operatorname{vol}\left(I_{k}\right)$. Since from hypothesis $f$ is unbounded in $\xi_{k}$, for every $M>0$, there exists $\xi_{k}^{\prime} \in I_{k}$ such that $\left\|f\left(\xi_{k}\right)-f\left(\xi_{k}^{\prime}\right)\right\|>M$ - that is, $\left\|f\left(\xi_{k}\right)-f\left(\xi_{k}^{\prime}\right)\right\|$ is obviously unbounded, which implies in the divergence of the Riemann sums, hence $f$ is non-Riemann integrable.

Definition A.7.10 (Riemann integral). The Riemann integral of real valued maps is an $\mathbf{R}$-linear map $\int: \mathcal{R} \rightarrow \mathbf{R}$ defined by mapping any $f \in \mathcal{R}(I)$ to

$$
\int_{I} f(x) \mathrm{d} x:=\lim _{\operatorname{mesh}(P) \rightarrow 0} \sigma(f, P, \xi)
$$

## Sets of Lebesgue Measure Zero

Definition A.7.11 (Set of Lebesgue measure zero). A set $E \subseteq \mathbf{R}^{n}$ is said be of Lebesgue measure zero if for every $\varepsilon>0$ there exists a countable open cover $\mathcal{U}$ of $E$ by $n$-dimensional open intervals whose total volume $\sum_{I \in \mathcal{U}} \operatorname{vol} \mathrm{Cl} I$ does not exceed $\varepsilon$.
Corollary A.7.12. A compact subset $E \subseteq \mathbf{R}^{n}$ is of measure zero if and only if, for all $\varepsilon$, there exists a finite open cover $\mathcal{U}$ of $E$ by open $n$-dimensional intervals such that $\sum_{I \in \mathcal{U}} \operatorname{volCl} I \leqslant \varepsilon$.
${ }^{1}$ For the time being, we have only defined the case where $E$ is an interval, but I'm already generalizing the notation for its uses in the following subsections.

Proof. If $E$ satisfies the last property, then clearly $E$ is a set of measure zero. On the other hand, if we assume that $E$ is of measure zero, given any $\varepsilon>0$, let $C$ be a countable cover of $E$ for which $\sum_{I \in C} \operatorname{volCl} I \leqslant \varepsilon$. Since $E$ is compact, there exists a finite subcover $\mathcal{U} \subseteq C$ of $E$, and since $\sum_{I \in \mathcal{U}} \operatorname{volCl} I \leqslant \sum_{I \in C} \operatorname{volCl} I \leqslant \varepsilon$, thus $\mathcal{U}$ is the wanted finite cover.

Lemma A.7.13. The following are properties of sets of measure zero:
(a) A subset of a set of measure zero is of measure zero.
(b) The countable union of sets of measure zero is of measure zero.
(c) A countable set is of measure zero.
(d) A non-degenerate interval is not a set of measure zero.

Proof. (a) Let $E \subseteq \mathbf{R}^{n}$ be a set of measure zero and $A \subseteq E$ be a subset. If $\mathcal{U}$ is a closed cover by intervals satisfying the measure zero condition, then in particular $\mathcal{U}$ covers $A$ therefore $A$ is of zero measure.
(b) Let $\left\{E_{j} \subseteq \mathbf{R}^{n}\right\}_{j \in J}$ be a countable collection of sets of measure zero, and let $\left\{\mathcal{U}_{j}\right\}_{j \in J}$ be a collection where $\mathcal{U}_{j}$ is the corresponding closed cover by intervals for $E_{j}$. Notice that the countable union $E:=\bigcup_{j \in J} E_{j}$ can be covered by $\mathcal{U}:=\bigcup_{j \in J} \mathcal{U}_{j}$, therefore, since the union of countable collections is countable, it follows that $\mathcal{U}$ is a countable cover for $E$ which satisfies the wanted property.
(c) We initially consider a single point in space. Notice that, for any given $\varepsilon$, there exists a closed interval (for instance, one could choose an interval of equal sides containing the point, whose sides have length less than $\varepsilon^{1 / n}$ ), whose volume is less than $\varepsilon$, containing the given point - that is, this one interval is sufficient to cover the point. We conclude that a singleton is of measure zero. Using the last item, we find that a countable set is of measure zero.
(d) Let $I=[a, b] \subseteq \mathbf{R}^{n}$ be a non-degenerate interval. Since $\mathbf{R}^{n}$ is Lindelöf, every cover of $I$ has a finite subcover so, we can proceed by induction on the cardinality $m \in \mathbf{N}$ of the open cover. For $m=1$, let $(\alpha, \beta) \subseteq \mathbf{R}^{n}$ be an open interval covering $I$. Notice that every $x \in I$ is such that $a_{j} \leqslant x_{j} \leqslant b_{j}$, for all $1 \leqslant j \leqslant n$, then, since $x$ must lie at $(\alpha, \beta)$, we necessarily have $\alpha_{j}<a \leqslant x_{j} \leqslant b<\beta_{j}$ so that $b_{j}-a_{j}<\beta_{j}-\alpha_{j}$ and hence $\operatorname{vol} I<\operatorname{vol}[\alpha, \beta]$. For the hypothesis of induction, suppose the proposition holds for a cover of cardinality $n-1 \in \mathbf{N}_{>1}$. Let $\left\{\left(\alpha^{i}, \beta^{i}\right) \subseteq \mathbf{R}^{n}\right\}_{i=1}^{m}$ be a cover of $I$ by open intervals. Let $1 \leqslant k \leqslant n$ be such that $a \in\left(\alpha^{k}, \beta^{k}\right)$, that is, $\alpha_{k}<a_{j}<\beta_{j}^{k}$. If for some index $1 \leqslant j_{0} \leqslant n$ we have $b_{j_{0}}>\beta_{j_{0}}^{k}$, we define the point $\beta^{\prime}$ to be such that, if $\beta_{j}^{k}<b_{j}$ then $\beta_{j}^{\prime}=\beta_{j}^{k}$, otherwise, if $\beta_{j}^{k} \geqslant b_{j}$, we let $a_{j}<\beta_{j}^{\prime}<b_{j}$ — that is, we constructed a point so that the closed interval $\left[\beta^{\prime}, b\right]$ is non-degenerate. From the hypothesis of induction, every cover of $\left[\beta^{\prime}, b\right]$ with cardinality $m-1$ has total volume strictly greater than $\operatorname{vol}\left[\beta^{\prime}, b\right]$. In particular, since $\left[\beta^{\prime}, b\right] \subset[a, b]$ then $\left\{\left(\alpha^{i}, \beta^{i}\right)\right\}_{i=1}^{m}$ is a cover of $\left[\beta^{\prime}, b\right]$, notice that $\left[\beta^{\prime}, b\right] \cap\left(\alpha^{k}, \beta^{k}\right)=\varnothing$, thus the cover $\left\{\left(\alpha^{i}, \beta^{i}\right)\right\}_{i=1, i \neq k}^{m}$ is a cover of $\left[\beta^{\prime}, b\right]$ with cardinality $n-1$, hence $\operatorname{vol}\left[\beta^{\prime}, b\right]<\sum_{i=1, i \neq k}^{m} \operatorname{vol}\left[\alpha^{i}, \beta^{i}\right]$. Then we find
that

$$
\begin{aligned}
b_{j}-a_{j} & <\left(b_{j}-\beta_{j}^{\prime}\right)+\left(\beta_{j}^{\prime}-a_{j}\right) \\
& \leqslant\left(b_{j}-\beta_{j}^{\prime}\right)+\left(\beta_{j}^{k}-a_{j}\right) \\
& <\left(b_{j}-\beta_{j}^{\prime}\right)+\left(\beta_{j}^{k}-\alpha_{j}^{k}\right) .
\end{aligned}
$$

Therefore we conclude that

$$
\begin{aligned}
\operatorname{vol}[a, b] & <\operatorname{vol}\left[\beta^{\prime}, b\right]+\operatorname{vol}\left[\alpha^{k}, \beta^{k}\right] \\
& <\sum_{i=1, i \neq k}^{m} \operatorname{vol}\left[\alpha^{i}, \beta^{i}\right]+\operatorname{vol}\left[\alpha^{k}, \beta^{k}\right] \\
& =\sum_{i=1}^{m} \operatorname{vol}\left[\alpha^{i}, \beta^{i}\right]
\end{aligned}
$$

which proves that the proposition is true for all $m \in \mathbf{N}$.

Example A.7.14. Let $f: I \rightarrow \mathbf{R}$ be a continuous map on the interval $I \subseteq \mathbf{R}^{n-1}$. The graph of $f$ is a $n$-dimensional set of Lebesgue measure zero.

Proof. Let $\Gamma$ denote the graph of $f$ over $I$. To see this, notice that since $I$ is closed, then $f$ is uniformly continuous on $I$. For any $\varepsilon>0$, let $\delta>0$ be such that $\|f(x)-f(y)\|_{\mathbf{R}}<\varepsilon$ for every $x, y \in I$ such that $\|x-y\|_{\mathbf{R}^{n-1}}<\delta$. Let $P$ be a partition of $I$ with mesh $P<\delta$. For any point $x_{0} \in R$ of each $R \in P$, we have an induced interval $R^{\prime}:=R \times\left[f\left(x_{0}\right)-\varepsilon, f\left(x_{0}\right)+\varepsilon\right]$ which is such that $\Gamma \subseteq R^{\prime}$, since we have an oscillation $\omega(f, R)<\varepsilon$ from the construction of $P$. Let $P^{\prime}$ be the collection of induced intervals $R^{\prime}$ of the partition $P$. From the last observation we have that $P^{\prime}$ is a closed cover of $\Gamma$ by closed $n$-dimensional intervals. Moreover, $\sum_{R^{\prime} \in P^{\prime}} \operatorname{vol} R^{\prime}=\sum_{R \in P} 2 \varepsilon \operatorname{vol} R=2 \varepsilon \operatorname{vol} I$ - therefore $\Gamma$ indeed is of Lebesgue measure zero

Notation A.7.15. Given a set $X$ and a property $P$, we say that $P$ holds almost everywhere on $X$ if the subset $A \subseteq X$, such that $P$ is not true, is a set of measure zero.

Theorem A.7.16 (Lebesgue's criterion). A map $f: I \rightarrow \mathbf{R}$ is Riemann integrable if and only if $f$ is bounded on $I$ and $f$ is continuous almost everywhere on $I$.

Proof. (Necessity) Let $f$ be Riemann integrable, then, from Proposition A.7.9, $f$ is bounded on $I$. For the sake of contradiction, let $E \subseteq I$ be the set composed of the points of discontinuity of $f$, we'll suppose that $E$ doesn't have measure zero. Notice that if $x \in E$, then there exists $n \in \mathbf{N}$ for which $\omega(f, x) \geqslant 1 / n$ - that is, $f$ does not converge to a value in $x$. We can then define $E_{n}:=\{x \in I: \omega(f, x) \geqslant 1 / n\}$ for every $n \in \mathbf{N}$ so that $E_{n} \subseteq E$ and thus $E=\bigcup_{n \in \mathbf{N}} E_{n}$. Since $E$ isn't of measure zero from assumption, it follows that there necessarily exists at least one $n_{0} \in \mathbf{N}$ such that $E_{n_{0}}$ isn't of measure zero.

Let $P$ be a partition of $I$, we'll consider two subsets of this partition:

$$
A:=\left\{R \in P: R \cap E_{n_{0}} \neq \varnothing \text { and } \omega(f, R) \geqslant 1 /\left(2 n_{0}\right)\right\},
$$

and $B:=P \backslash A$. Since $P$ partitions the interval, for any $x \in E_{n_{0}}$, there exists $R \in P$ such that $x \in \operatorname{Int} R$ or $x \in \partial R$ - we now analyse both cases:

- In the case where $x$ is an interior point, since $\omega(f, x) \geqslant 1 / n_{0}$, it follows that $R \in A$.
- Otherwise, if $x$ is a boundary point, then there actually exists at least another $R^{\prime} \in P$ such that $x \in \partial R \cap \partial R^{\prime}$. Suppose, for the sake of contradiction, that every interval $R^{\prime} \in P$ containing $x$ as a boundary point is such that $f$ has an oscillation $\omega\left(f, R^{\prime}\right)<1 /\left(2 n_{0}\right)$, then, if we take any ball $B_{x}(r) \cap I$, neighbourhood of $x$, we find that, $\omega\left(f, B_{x}(r)\right)<1 /\left(2 n_{0}\right)+1 /\left(2 n_{0}\right)=1 / n_{0}$, that is, the limit $\omega(f, x)<1 / n_{0}$, which is a contradiction to the assumption that $x \in E_{n}$ - thus there must exist $R^{\prime} \in P$ with $x \in \partial R^{\prime}$ such that $\omega\left(f, R^{\prime}\right) \geqslant 1 /\left(2 n_{0}\right)$ so that $R^{\prime} \in A$.

This implies that $A$ covers the interval $E_{n_{0}}$ by closed intervals and, by assumption, $\sum_{R \in A} \operatorname{vol} R>\operatorname{vol} E_{n_{0}}$.

We are now going to consider any two distinct choices of distinguished points $\xi$ and $\xi^{\prime}$ of $P$ such that, if $\xi_{j}$ and $\xi_{j}^{\prime}$ are elements of a common interval of $B$ then $\xi_{j}=\xi_{j}^{\prime}$, and if $\xi_{j}$ and $\xi_{j}^{\prime}$ are elements common to an interval of $A$, we choose $\xi_{j}$ and $\xi_{j}$ to be any points such that $f\left(\xi_{j}\right)-f\left(\xi_{j}^{\prime}\right)>1 /\left(3 n_{0}\right)$ - which is always possible from the construction of $A$. Notice that we have

$$
\begin{aligned}
\left\|\sigma(f, P, \xi)-\sigma\left(f, P, \xi^{\prime}\right)\right\| & =\left\|\sum_{R_{j} \in A}\left(f\left(\xi_{j}\right)-f\left(\xi_{j}^{\prime}\right)\right) \operatorname{vol} R_{j}\right\| \\
& >\frac{1}{3 n_{0}} \sum_{R_{j} \in A} \operatorname{vol} R_{j} \\
& >\frac{1}{3 n_{0}} \operatorname{vol} E_{n_{0}}>0
\end{aligned}
$$

and therefore the limit $\lim _{\operatorname{mesh}(P) \rightarrow 0} \sigma(f, P, \xi)$ does not converge whenever $E_{n_{0}}$ is of measure zero - that is, if the set of discontinuities of $f$ isn't of measure zero, then $f$ is non-Riemann integrable.
(Sufficiency) Suppose that $f$ is bounded and continuous almost everywhere on I. Let $\varepsilon>0$ be any constant and consider the subset $E_{\varepsilon}:=\{x \in I: \omega(f, x) \geqslant \varepsilon\}$ of discontinuous points of $f$. If $E$ is the set of points of discontinuity of $f$, then by hypothesis $E$ is of measure zero - since $E_{\varepsilon} \subseteq E$, then $E_{\varepsilon}$ is also of measure zero. Moreover, $E_{\varepsilon}$ is necessarily closed, which implies in $E_{\varepsilon}$ compact. Let $\left\{I_{j}\right\}_{j=1}^{k}$ be a finite open cover of $E_{\varepsilon}$ by open intervals $I_{j} \subseteq \mathbf{R}^{n}$ such that $\sum_{i=1}^{k} \operatorname{volCl} I<\varepsilon$ (recall that this is possible because of Corollary A.7.12).

We now define the $\lambda$-dilations of intervals of the cover $\left\{I_{j}\right\}_{j=1}^{k}$ but with centre unchanged - that is, if $I_{j}=(x, y)$, then the dilation $I_{j}^{\lambda}:=(z, w)$ must be such that $w_{i}-z_{i}=\lambda^{1 / n}\left(y_{i}-x_{i}\right)$ and $w_{i}+z_{i}=y_{i}+x_{i}$, for all $1 \leqslant i \leqslant n$. Solving such system we
obtain $z_{i}=\frac{1}{2}\left[x_{i}\left(1+\lambda^{1 / n}\right)+y_{i}\left(1-\lambda^{1 / n}\right)\right]$ and $w_{i}=\frac{1}{2}\left[x_{i}\left(1-\lambda^{1 / n}\right)+y_{i}\left(1+\lambda^{1 / n}\right)\right]$. This way we have

$$
\sum_{j=1}^{k} \operatorname{volCl} I_{j}^{\lambda}=\lambda^{n} \sum_{j=1}^{k} \operatorname{volCl} I_{j}<\lambda^{n} \varepsilon
$$

We'll define the sets $C_{\lambda}:=\left(\bigcup_{j=1}^{k} I_{j}^{\lambda}\right) \cap I$ - moreover, $d$ will be the minimum distance between the boundaries of $C_{2}$ and $C_{3}$.

From construction, we have $E_{\varepsilon} \subseteq \operatorname{Int} C_{2}$, now if we consider the compact set $K:=$ $I \backslash\left(\operatorname{Int} C_{2}\right)$, then, for every $x \in K$, we have $\omega(f, x)<\varepsilon$. Using Theorem A.2.33, we find that there exists a $\delta>0$ such that $\|f(x)-f(y)\|<2 \varepsilon$, for all $x, y \in K$ with $\|x-y\|<\delta$. We'll now finally show the convergence of the Riemann sums. Let $P^{\prime}:=\left\{R_{\alpha}^{\prime}\right\}_{\alpha \in A}$ and $P^{\prime \prime}:=\left\{R_{\beta}^{\prime \prime}\right\}_{\beta \in B}$ be any two partitions of $I$ with $\operatorname{mesh}\left(P^{\prime}\right), \operatorname{mesh}\left(P^{\prime \prime}\right)<\min (d, \delta)$. Define $P:=\left\{R_{(\alpha, \beta)}:=R_{\alpha}^{\prime} \cap R_{\beta}^{\prime \prime}\right\}_{(\alpha, \beta) \in A \times B,}$, which clearly is a partition of $I$. Since we are dealing with partitions, for every $R_{\alpha}^{\prime} \in P^{\prime}$ we have $R_{\alpha}^{\prime}=\bigcup_{\beta \in B} R_{(\alpha, \beta)}$ - thus $\operatorname{vol} R_{\alpha}^{\prime}=\sum_{\beta \in B} \operatorname{vol} R_{(\alpha, \beta)}$. Let $\xi^{\prime}$ and $\xi$ be any two distinguished points of the partitions $P^{\prime}$ and $P$, respectively, then

$$
\begin{aligned}
\left\|\sigma\left(f, P^{\prime}, \xi^{\prime}\right)-\sigma(f, P, \xi)\right\|= & \left\|\sum_{(\alpha, \beta) \in A \times B}\left(f\left(\xi_{\alpha}^{\prime}\right)-f\left(\xi_{(\alpha, \beta)}\right)\right) \operatorname{vol} R_{(\alpha, \beta)}\right\| \\
\leqslant & \sum_{(\alpha, \beta) \in A^{\prime} \times B^{\prime}}\left\|f\left(\xi_{\alpha}^{\prime}\right)-f\left(\xi_{(\alpha, \beta)}\right)\right\| \operatorname{vol} R_{(\alpha, \beta)} \\
& +\sum_{(\alpha, \beta)(A \times B) \backslash\left(A^{\prime} \times B^{\prime}\right)}\left\|f\left(\xi_{\alpha}^{\prime}\right)-f\left(\xi_{(\alpha, \beta)}\right)\right\| \operatorname{vol} R_{(\alpha, \beta)},
\end{aligned}
$$

where we define

$$
A^{\prime} \times B^{\prime}:=\left\{(\alpha, \beta) \in A \times B: R_{\alpha}^{\prime} \subseteq C_{3}\right\}
$$

Moreover, since the diameter $d\left(R_{\alpha}^{\prime}\right)<d$, then for every $R_{\alpha}^{\prime}$ not entirely contained in $C_{3}$, the intersection with any $R_{\beta}^{\prime \prime}$, namely $R_{(\alpha, \beta)}$, cannot lie in the interior of $C_{2}$ - that is, for all $(\alpha, \beta) \in(A \times B) \backslash\left(A^{\prime} \times B^{\prime}\right)$ we'll have $R_{(\alpha, \beta)} \subseteq K$ and surely $\xi_{\alpha}^{\prime}, \xi_{(\alpha, \beta)} \in K$, which satisfy $\left\|f\left(\xi_{\alpha}^{\prime}\right)-f\left(\xi_{(\alpha, \beta)}\right)\right\|<2 \varepsilon$ since mesh $P^{\prime}<\delta$. Assuming $f$ is bounded by $M>0$ in $I$, that is $\|f\| \leqslant M$, then $\left\|f\left(\xi_{\alpha}^{\prime}\right)-f\left(\xi_{(\alpha, \beta)}\right)\right\|<2 M$ and thus

$$
\begin{aligned}
\left\|\sigma\left(f, P^{\prime}, \xi^{\prime}\right)-\sigma(f, P, \xi)\right\| & \leqslant \sum_{(\alpha, \beta) \in A^{\prime} \times B^{\prime}} 2 M \operatorname{vol} R_{(\alpha, \beta)}+\sum_{(\alpha, \beta) \in(A \times B) \backslash\left(A^{\prime} \times B^{\prime}\right)} 2 \varepsilon \operatorname{vol} R_{(\alpha, \beta)} \\
& \leqslant 2 M\left(3^{n} \varepsilon\right)+2 \varepsilon \operatorname{vol} I \\
& =2 \varepsilon\left(3^{n} M+\operatorname{vol} I\right) .
\end{aligned}
$$

Therefore, since the same construction is applicable for $P^{\prime \prime}$, that is, $\| \sigma\left(f, P^{\prime \prime}, \xi^{\prime \prime \prime}\right)-$ $\sigma(f, P, \xi) \| \leqslant 2 \varepsilon\left(3^{n} M+\operatorname{vol} I\right)$, thus

$$
\left\|\sigma\left(f, P^{\prime}, \xi^{\prime}\right)-\sigma\left(f, P^{\prime \prime}, \xi^{\prime \prime}\right)\right\| \leqslant 4 \varepsilon\left(3^{n} M+\operatorname{vol} I\right)
$$

Since $P^{\prime}$ and $P^{\prime \prime}$ where chosen arbitrarily, it follows that the sequence of Riemann sums converge by the Cauchy criterion and thus $f$ is Riemann integrable from definition. $\quad \square$

## Integrating Over Sets

Definition A.7.17. A set $E \subseteq \mathbf{R}^{n}$ is said to be admissible if it is bounded and has measure zero boundary.

Corollary A.7.18 (Operations on admissible sets). The finite union or intersection of admissible sets is admissible, and the difference of admissible sets is admissible.

Notation A.7.19. We denote the characteristic map of a given set $E$ to be $\chi_{E}: E \rightarrow\{0,1\}$, where $\chi_{E}(x):=1$ if $x \in E$, and $\chi_{E}(x):=0$ if $x \notin E$. Moreover, given any map $f: E \rightarrow \mathbf{R}$, we define the $\operatorname{map} f_{\chi_{E}}: \mathbf{R}^{n} \rightarrow \mathbf{R}$ as $\left.f_{\chi_{E}}\right|_{E}:=f$ and $f_{\chi_{E}}$ is zero everywhere else.

Definition A.7.20 (Riemann integral over a set). Let $E \subseteq \mathbf{R}^{n}$ be an admissible set and $f: I \rightarrow \mathbf{R}$ for some $I \supseteq E$, then we define the integral of $f$ over the set $E$ as

$$
\int_{E} f(x) \mathrm{d} x:=\int_{I} f_{\chi_{E}}(x) \mathrm{d} x
$$

Lemma A.7.21 (The integral is well-defined). Let $f: E \rightarrow \mathbf{R}$ be a map. The integral of $f$ over $E$ is independent of the choice of the interval containing $E$.

Proof. Let $I, I^{\prime} \subseteq \mathbf{R}^{n}$ be any two intervals containing $E$. Define an auxiliary interval $I_{0}:=I \cap I^{\prime}$. It follows from construction that every point of discontinuity of $f_{\chi_{E}}$ is contained in $I_{0}$. From Theorem A.7.16 we see that, if the collection of points of discontinuity of $f_{\chi_{E}}$ is not of measure zero, then both integrals of $f_{\chi_{E}}$ over $I$ and $I^{\prime}$ fail to exist - while, if the collection is of measure zero, both integrals exist simultaneously.

Suppose that the points of discontinuity form a set of measure zero. Given any $\varepsilon>0$, we consider a partition $P_{0}$ of $I$ with mesh $P_{0}<\varepsilon$. Define partitions $P$ and $P^{\prime}$ of $I$ and $I^{\prime}$, respectively, such that $P_{0} \subseteq P \cap P^{\prime}$ - that is, inside $I_{0}$, they share the exact same collection of intervals as $P_{0}$. Given any distinguished points $\xi$ and $\xi^{\prime}$ of $P$ and $P^{\prime}$, respectively, we have that $f_{\chi_{E}}(x)=0$ for every $x \in \xi \cup \xi^{\prime}$ such that $x \notin I_{0}$, therefore the Riemann sums of $f_{\chi_{E}}$ under the partitions $P$ and $P^{\prime}$ are always reduced to Riemann sums of $f_{\chi_{E}}$ under the partition $P_{0}$ and the corresponding distinguished points. This implies that the limit of the Riemann sums are equal for both partitions and therefore we have

$$
\int_{I} f_{\chi_{E}}(x) \mathrm{d} x=\int_{I^{\prime}} f_{\chi_{E}}(x) \mathrm{d} x
$$

Corollary A.7.22. Let $E \subseteq \mathbf{R}^{n}$ be an admissible set and $f: E \rightarrow \mathbf{R}$. Then, $f$ is Riemann integrable if and only if is bounded and continuous at almost all points of $E$.

Proof. Notice that the discontinuities of the corresponding map $f_{\chi_{E}}$ are those of $f$ and perhaps a collection of points of $\partial E$, where $\chi_{E}$ changes its value. Since $\partial E$ is of measure zero, it doesn't interfere in the use of Lebesgue's criterion and thus the proposition follows.

## Jordan Measure

Definition A.7.23 (Jordan measure). Given a bounded set $E \subseteq \mathbf{R}$, we define the Jordan measure of $E$ as the map

$$
\mu(E):=\int_{E} \mathrm{~d} x .
$$

Notice that the integral over $E$ only exists for admissible sets, thus the Jordan measure $\mu$ is only defined for admissible sets.

Definition A.7.24 (Jordan's measure zero sets). A set $E \subseteq \mathbf{R}^{n}$ is said to be of Jordan measure zero (or of content zero) if for every $\varepsilon>0$ there exists a finite open cover by intervals whose total volume is less than $\varepsilon$.

Definition A.7.25 (Jordan-measurable sets). A set $E \subseteq \mathbf{R}^{n}$ is said to be Jordan-measurable if it is bounded and $\partial E$ is a set of Jordan measure zero.

## Riemann Integral Properties

Corollary A.7.26. The collection of Riemann integrable maps $E \rightarrow \mathbf{R}$ for an $\mathbf{R}$-vector space. Moreover, the Riemann integral is an $\mathbf{R}$-linear functional $\int_{E}: \mathcal{R}(E) \rightarrow \mathbf{R}$ of the dual space $\mathcal{R}(E)^{*}$.

Proposition A.7.27. Let $E, S \subseteq \mathbf{R}^{n}$ be admissible sets and $f: E \cup S \rightarrow \mathbf{R}$. Then
(a) $f$ is Riemann integrable over $E \cup S$ if and only if it's integrable over both $E$ and $S$ simultaneously. Moreover, if such condition is met, the $f$ is also Riemann integrable over $E \cap S$.
(b) If $f$ is Riemann integrable over $E \cup S$ and $\mu(E \cap S)=0$, then

$$
\int_{E \cup S} f(x) \mathrm{d} x=\int_{E} f(x) \mathrm{d} x+\int_{S} f(x) \mathrm{d} x
$$

Proof. (a) If $f$ is Riemann integrable over $E \cup S$, then $f$ is bounded and continuous at almost all points of $E \cup S$, which implies that $f$ is Riemann integrable over both $E$ and $S$. Since $E \cap S \subseteq E \cup S$ then $f$ is also Riemann integrable over the intersection.
(b) Notice that $\chi_{E \cup S}(x)=\chi_{E}(x)+\chi_{S}(x)-\chi_{E \cap S}(x)$, therefore by definition -if $I \supseteq E \cup S$ is an interval, then

$$
\begin{aligned}
\int_{E \cup S} f(x) \mathrm{d} x & =\int_{I} f_{\chi_{E \cup S}}(x) \mathrm{d} x \\
& =\int_{I} f_{\chi_{E}}(x)+f_{\chi S}(x)-f_{\chi \text { EnS }}(x) \mathrm{d} x \\
& =\int_{I} f_{\chi_{E}}(x) \mathrm{d} x+\int_{I} f_{\chi_{S}}(x) \mathrm{d} x-\int_{I} f_{\chi_{\text {EnS }}}(x) \mathrm{d} x \\
& =\int_{I} f_{\chi_{E}}(x) \mathrm{d} x+\int_{I} f_{\chi_{S}}(x) \mathrm{d} x
\end{aligned}
$$

where we used the hypothesis that $\mu(E \cap S)=0$ to conclude that

$$
\int_{I} f_{\chi \mathrm{E} \cap \mathrm{~S}}(x) \mathrm{d} x=\int_{E \cap S} f(x) \mathrm{d} x=0 .
$$

Proposition A.7.28 (Estimate). Let $f \in \mathcal{R}(E)$, where $E \subseteq \mathbf{R}^{n}$ is an admissible set, then $\|f\| \in \mathcal{R}(E)$ and also

$$
\left\|\int_{E} f(x) \mathrm{d} x\right\| \leqslant \int_{E}\|f(x)\| \mathrm{d} x
$$

Proof. Notice that if $f$ is Riemann integrable, then $f$ is bounded and and continuous at almost all points of $E$, thus surely $\|f\|$ is bounded and continuous - hence Riemann integrable. Moreover, since $f(x) \leqslant\|f(x)\|$ for all $x \in E$, it follows that, for any given partition $P$ and distinguished points $\chi$, we have the inequality $\sigma(f, P, \chi) \leqslant \sigma(\|f\|, P, \chi)$, thus the integral inequality holds by taking the limit mesh $P \rightarrow 0$.

Proposition A.7.29 (Non-negative real valued maps). Let $f: E \rightarrow \mathbf{R}$ be a Riemann integrable map over the admissible set $E \subseteq \mathbf{R}^{n}$. If $f$ is non-negative over $E$, that is, $f(x) \geqslant 0$ for all $x \in E$, then

$$
\int_{E} f(x) \mathrm{d} x \geqslant 0
$$

Proof. Just observe that for any partition $P$ and distinguished points $\chi$, we have $\sigma(f, P, \chi) \geqslant 0$, thus $\lim _{\text {mesh }} P \rightarrow 0 \sigma(f, P, \chi) \geqslant 0$.

Corollary A.7.30. Let $E \subseteq R^{n}$ be an admissible set and $f, g: E \rightarrow \mathbf{R}$ be two given maps. The following are miscellaneous immediate implications of Corollary A.7.30:
(a) If both $f$ and $g$ are Riemann integrable, and $f(x) \leqslant g(x)$ for all $x \in E$, then

$$
\int_{E} f(x) \mathrm{d} x \leqslant \int_{E} g(x) \mathrm{d} x
$$

(b) If $f$ is Riemann integrable over $E$ and, for some constants $m, M \in \mathbf{R}$ we have $m \leqslant f(x) \leqslant M$ for all $x \in E$, then

$$
m \mu(E) \leqslant \int_{E} f(x) \mathrm{d} x \leqslant M \mu(E)
$$

(c) If $f$ is Riemann integrable over $E$, define constants $m:=\inf _{x \in E} f(x)$ and $M:=$ $\sup _{x \in E} f(x)$. Then, there exist a constant $m \leqslant \omega \leqslant M$ for which

$$
\int_{E} f(x) \mathrm{d} x=\theta \mu(E)
$$

(d) If additionally $E$ is connected and $f$ is continuous, then $f$ is Riemann integrable over $E$ and there exists $y \in E$ for which

$$
\int_{E} f(x) \mathrm{d} x=f(y) \mu(E)
$$

(e) If $f$ and $g$ are Riemann integrable over $E$ and, for some constants $m, M \in \mathbf{R}$ we have $m \leqslant f(x) \leqslant M$ for all $x \in E$, and $g(x) \geqslant 0$ for all $x \in E$, then

$$
m \int_{E} g(x) \mathrm{d} x \leqslant \int_{E} f(x) g(x) \mathrm{d} x \leqslant M \int_{E} g(x) \mathrm{d} x
$$

Lemma A.7.31. Let $f: E \rightarrow \mathbf{R}$ be a non-negative map that is Riemann integrable over $E$, where $E$ is an admissible set. If $\int_{E} f(x) \mathrm{d} x=0$, then $f(x)=0$ at almost all points of $E$.

Proof. Let $x_{0} \in E$ be any point of continuity of $f$ and, for the sake of contradiction, assume that $f\left(x_{0}\right)>0$ and consider a constant $c>0$ with $f\left(x_{0}\right) \geqslant c>0$. Let $U \subseteq E$ be any neighbourhood of $x_{0}$ for which $f(x) \geqslant c>0$ for every $x \in U$. Then we obtain the following contradiction - where $I \supseteq E$ is an interval:
$\int_{E} f(x) \mathrm{d} x=\int_{I} f_{\chi_{E}}(x) \mathrm{d} x=\int_{U} f_{\chi_{E}}(x) \mathrm{d} x+\int_{I \backslash U} f_{\chi_{E}}(x) \mathrm{d} x \geqslant \int_{U} f_{\chi_{E}}(x) \mathrm{d} x \geqslant c \mu(U)>0$.
This cannot be true, thus $f\left(x_{0}\right)=0$ for every point $x_{0} \in E$ where $f$ is continuous. Since $f$ is continuous almost everywhere, it follows that $f$ is zero almost everywhere. $\quad$

Corollary A.7.32. Let $\sim$ be the equivalence relation on $\mathcal{R}(E)$ as follows: $f \sim g$ if and only if the collection of points where the map $f-g$ is non-zero forms a set of Lebesgue measure zero. Then the map $\|-\|: \mathcal{R}(E) / \sim \rightarrow \mathbf{R}$ defined by

$$
\|f\|:=\int_{E}\|f(x)\|_{\mathbf{R}^{n}} \mathrm{~d} x
$$

is a norm on the vector space $\mathcal{R}(E) / \sim$.

## A. 8 Fubini's Theorem

## Upper and Lower Darboux Integrals

In order to proceed to the theorem of Fubini type, we first need to define the concept of Darboux lower and upper integrals.

Definition A.8.1 (Darboux sums). Let $f: I \rightarrow \mathbf{R}$ be a map defined on the interval $I \subseteq \mathbf{R}^{n}$. If $P:=\left\{I_{j}\right\}_{j \in J}$ is a partition of $I$ by intervals, then we define the following

- The lower Darboux sum of $f$ over the interval $I$ corresponding to the partition $P$ is

$$
s(f, P):=\sum_{j \in J} \operatorname{vol}\left(I_{j}\right) \inf _{x \in I_{j}} f(x)
$$

- The upper Darboux sum of $f$ over the interval $I$ corresponding to the partition $P$ is

$$
S(f, P):=\sum_{j \in J} \operatorname{vol}\left(I_{j}\right) \sup _{x \in I_{j}} f(x) .
$$

Definition A.8.2 (Upper and lower Darboux integrals). Let $f: I \rightarrow \mathbf{R}$ be a map over the interval $I \subseteq \mathbf{R}^{n}$. We define:

- The lower Darboux integral of $f$ is $\underline{\mathcal{J}}:=\sup _{P} s(f, P)$, where the $P$ goes through all partitions of $I$.
- The upper Darboux integral of $f$ is $\underline{\mathcal{J}}:=\inf _{P} S(f, P)$, where the $P$ goes through all partitions of $I$.

Theorem A.8.3 (Darboux criterion). A map $f: I \rightarrow \mathbf{R}$, where $I \subseteq \mathbf{R}^{n}$ is an interval, is integrable over $I$ if and only if $f$ is bounded on $I$ and the upper and lower Darboux integrals agree over $I$.

## Add proof of Darboux criterion later

## Fubini

Theorem A.8.4. Let $X \times Y \subseteq \mathbf{R}^{m+n}$ be an interval, and $f: X \times Y \rightarrow \mathbf{R}$ be a Riemann integrable map over $X \times Y$. Then we have the following equality:

$$
\begin{equation*}
\int_{X \times Y} f(x, y) \mathrm{d} x \mathrm{~d} y=\int_{X}\left(\int_{Y} f(x, y) \mathrm{d} y\right) \mathrm{d} x=\int_{Y}\left(\int_{X} f(x, y) \mathrm{d} x\right) \mathrm{d} y \tag{A.45}
\end{equation*}
$$

Proof. In order to prove the theorem, we first construct the following maps:

- $F_{X}: X \rightarrow \mathbf{R}$ is defined as $F_{X}\left(x_{0}\right):=\int_{Y} f\left(x_{0}, y\right) \mathrm{d} x$ whenever such integral exists, in case not, we let $\mathcal{J}_{x_{0}} \leqslant F_{X}\left(x_{0}\right) \leqslant \overline{\mathcal{J}_{x_{0}}}$ be any element between the Darboux integrals of the map $\left.\overline{f( } x_{0},-\right): Y \rightarrow \mathbf{R}$ given by $y \mapsto f\left(x_{0}, y\right)$.
- $F_{Y}: Y \rightarrow \mathbf{R}$ is defined as $F_{Y}\left(y_{0}\right):=\int_{x} f\left(x, y_{0}\right) \mathrm{d} x$ whenever such integral exists, in case not, we let $\mathcal{J}_{y_{0}} \leqslant F_{Y}\left(y_{0}\right) \leqslant \overline{\mathcal{J}_{y_{0}}}$ be any element between the Darboux integrals of the map $f\left(-, y_{0}\right): X \rightarrow \mathbf{R}$ given by $x \mapsto f\left(x, y_{0}\right)$.

We'll carry the proof for the first equality, but, knowing the construction of $F_{Y}$, the idea for the second equality is the exact analogue.

Let $P:=\left\{X_{i}\right\}_{i \in I} \times\left\{Y_{j}\right\}_{j \in J}$ be any partition of the interval $X \times Y$. Notice that

$$
\begin{aligned}
s(f, P) & =\sum_{(i, j) \in I \times J} \inf _{(x, y) \in X_{i} \times y_{j}} f(x, y) \operatorname{vol}\left(X_{i} \times Y_{j}\right) \\
& =\sum_{(i, j) \in I \times J} \inf _{(x, y) \in X_{i} \times y_{j}} f(x, y) \operatorname{vol}\left(X_{i}\right) \operatorname{vol}\left(Y_{j}\right) \\
& \leqslant \sum_{i \in I} \inf _{x \in X_{i}}\left[\sum_{j \in J} f(x, y) \operatorname{vol}\left(Y_{j}\right)\right] \operatorname{vol} X_{i} \\
& \leqslant \sum_{i \in I} \inf _{x \in X_{i}}\left[\underline{\int_{Y}} f(x, y) \mathrm{d} y\right] \operatorname{vol} X_{i} \\
& \leqslant \sum_{i \in I} \inf _{x \in X_{i}} F_{X}(x) \operatorname{vol} X_{i} \\
& \leqslant \sum_{i \in I} \sup _{x \in X_{i}} F_{X}(x) \operatorname{vol} X_{i} \\
& \leqslant \sum_{i \in I} \sup _{x \in X_{i}}\left[\int_{X} f(x, y) \mathrm{d} x\right] \operatorname{vol} X_{i} \\
& \leqslant \sum_{i \in I} \sup _{x \in X_{i}}\left[\sum_{j \in J} f(x, y) \operatorname{vol} Y_{j}\right] \operatorname{vol} X_{i} \\
& \leqslant \sum_{(i, j) \in I \times J} f(x, y) \operatorname{vol}\left(X_{i} \times Y_{j}\right) \\
& =S(f, P)
\end{aligned}
$$

By hypothesis, $f$ is Riemann integrable over $X \times Y$, thus

$$
\lim _{\operatorname{mesh} P \rightarrow 0} s(f, P)=\lim _{\operatorname{mesh} P \rightarrow 0} S(f, P)=\int_{X \times Y} f(x, y) \mathrm{d} x \mathrm{~d} y
$$

therefore $F_{X}$ is Riemann integrable over $X$ and

$$
\int_{X \times Y} f(x, y) \mathrm{d} x \mathrm{~d} y=\int_{X} F_{X}(x) \mathrm{d} x
$$

Corollary A.8.5. Let $f: X \times Y \rightarrow \mathbf{R}$ be Riemann integrable. Then for almost all points $x_{0} \in X$ and almost all $y_{0} \in Y$ the integrals $\int_{Y} f\left(x_{0}, y\right) \mathrm{d} y$ and $\int_{X} f\left(x, y_{0}\right) \mathrm{d} x$ exist.
Proof. Since $s(f, P) \leqslant S(f, P)$ for all partitions $P$ of $X \times Y$, then $\underline{\int_{Y}} f\left(x_{0}, y\right) \mathrm{d} y \leqslant$ $\overline{\int_{X}} f\left(x_{0}, y\right) \mathrm{d} y$ for every $x_{0} \in X$. From Theorem A.8.4 we see that

$$
\int_{X}\left(\overline{\int_{Y}} f(x, y) \mathrm{d} y-\int_{\underline{Y}} f(x, y) \mathrm{d} y\right) \mathrm{d} x=0
$$

therefore by Lemma A.7.31 we conclude that the map $x_{0} \mapsto \bar{\int}_{Y} f\left(x_{0}, y\right) \mathrm{d} y-\underline{\int_{Y}} f\left(x_{0}, y\right) \mathrm{d} y$ equals zero for almost all $x_{0} \in X$ - that is, $\int_{Y} f\left(x_{0}, y\right) \mathrm{d} y$ exists for almost all $x_{0} \in X$. The exact same argument can be used to show this for $\int_{X} f\left(x, y_{0}\right) \mathrm{d} x$.

Corollary A.8.6 (Iterated Fubini's theorem). Let $I:=\prod_{j=1}^{n}\left[a_{j}, b_{j}\right] \subseteq \mathbf{R}^{n}$ be an interval and $f: I \rightarrow \mathbf{R}$ be a Riemann integrable map over $I$. Then

$$
\int_{I} f(x) \mathrm{d} x=\int_{a_{n}}^{b_{n}}\left(\int_{a_{n-1}}^{b_{n-1}}\left(\ldots\left(\int_{a_{1}}^{b_{1}} f(x, y) \mathrm{d} x_{1}\right) \ldots\right) \mathrm{d} x_{n-1}\right) \mathrm{d} x_{n}
$$

Moreover, the result of the integral is invariant under any permutation of the order of the sub-intervals $\left[a_{j}, b_{j}\right]$ of integration.

Proof. This is just a direct application of Theorem A.8.4, moreover, the second part of the proposition can be obtained by seeing that the second equality of Eq. (A.45) is nothing but a transposition on the set $\{X, Y\}$ - generalizing this for transpositions on $\left\{\left[a_{j}, b_{j}\right]\right\}_{j=1}^{n}$ imply that any permutation of the collection satisfies the integral equality.

## Measuring Volumes Between Graphs of Maps

Corollary A.8.7. Define $D \subseteq \mathbf{R}^{n-1}$ to be a bounded set and let maps $g_{1}, g_{2}: D \rightarrow \mathbf{R}$ such that $g_{1} \leqslant g_{2}$. Let $f: E \rightarrow \mathbf{R}$ be a Riemann integrable map over $E:=\left\{(x, y) \in \mathbf{R}^{n-1} \times \mathbf{R}\right.$ : $x \in D$ and $\left.g_{1}(x) \leqslant y \leqslant g_{2}(x)\right\}$. Then

$$
\int_{E} f(x, y) \mathrm{d} x \mathrm{~d} y=\int_{D}\left(\int_{g_{1}(x)}^{g_{2}(x)} f(x, y) \mathrm{d} y\right) \mathrm{d} x
$$

Proof. For each $x \in D$ define the set $E_{x}:=\left\{y \in \mathbf{R}: g_{1}(x) \leqslant y \leqslant g_{2}(x)\right\}$, and observe that, from construction, $\chi_{E}(x, y)=\chi_{D}(x) \chi_{E_{x}}(y)$. Given any interval $I:=I_{x} \times I_{y} \supseteq E$, we have

$$
\int_{E} f(x, y) \mathrm{d} x \mathrm{~d} y=\int_{I} f_{\mathcal{X}_{E}}(x, y) \mathrm{d} x \mathrm{~d} y=\int_{I_{x}}\left(\int_{I_{y}} f_{\mathcal{X}_{E}}(x, y) \mathrm{d} y\right) \mathrm{d} x
$$

Corollary A.8.8 (Measuring the set between the graph of continuous maps). Let $D \subseteq$ $\mathbf{R}^{n-1}$ be a Jordan-measurable set and $g_{1}, g_{2}: D \rightarrow \mathbf{R}$ be continuous maps such that $g_{1} \leqslant g_{2}$. Then the set $E:=\left\{(x, y) \in \mathbf{R}^{n-1} \times \mathbf{R}: x \in D\right.$ and $\left.g_{1}(x) \leqslant y \leqslant g_{2}(x)\right\} \subseteq \mathbf{R}^{n}$ is Jordan-measurable and its Jordan measure is

$$
\mu(E)=\int_{D} g_{2}(x)-g_{1}(x) \mathrm{d} x
$$

Proof. Denote by $\Gamma_{g_{1}}$ and $\Gamma_{g_{2}}$ the graphs of $g_{1}$ and $g_{2}$, respectively. By Example A.7.14, $\Gamma_{g_{1}}$ and $\Gamma_{g_{2}}$ are sets of Lebesgue measure zero. Define a set

$$
Z:=\partial D \times\left[\inf _{x \in D} g_{1}(x), \sup _{x \in D} g_{2}(x)\right] \subseteq \mathbf{R}^{n}
$$

and the length $\ell:=\sup _{x \in d} g_{2}(x)-\inf _{x \in D} g_{1}(x)$. Since $D$ is a Jordan-measurable set, it follows that $\partial D$ can be covered by a finite collection of open sets with total volume arbitrarily small - in particular, smaller than $\varepsilon / \ell$. In particular, it follows that $Z$ is a set of Jordan measure zero. Notice that $\partial E \subseteq \Gamma_{g_{1}} \cup \Gamma_{g_{2}} \cup Z$ and thus $\partial E$ is a set of Jordan measure zero - thus $E$ is Jordan-measurable. The measure formula for $E$ follows immediately by Corollary A.8.7

Corollary A.8.9. Let $E \subseteq \mathbf{R}^{n}$ be a Jordan-measurable set and $I:=I_{x} \times I_{y} \subseteq \mathbf{R}^{n-1} \times \mathbf{R}$ be an interval containing $E$. For each $y_{0} \in I_{y}$, define the section $E_{y_{0}}:=\left\{\left(x, y_{0}\right)\right\} \subseteq E$. Then, for almost all $y_{0} \in I_{y}$ the set $E_{y_{0}}$ is Jordan-measurable, and

$$
\mu(E)=\int_{I_{y}} F_{\mu}(y) \mathrm{d} y
$$

where the map $F_{\mu}: I_{y} \rightarrow \mathbf{R}$ is defined as $F_{\mu}(y):=\mu\left(E_{y}\right)$ if $E_{y}$ happens to be Jordanmeasurable, otherwise, we let $\int_{E_{y}} \mathrm{~d} y \leqslant F_{\mu}(y) \leqslant \overline{\int_{E_{y}}} \mathrm{~d} y$ be any value between the lower and upper Darboux integrals of the measure of $E_{y}$.

Corollary A.8.10 (Cavalieri's principle). Let $A$ and $B$ be two Jordan-measurable solids in $\mathbf{R}^{3}$. For every $z_{0} \in \mathbf{R}$, define the sets $A_{z_{0}}:=\left\{\left(x, y, z_{0}\right) \in A\right\}$ and $B_{z_{0}}:=\left\{\left(x, y, z_{0}\right) \in B\right\}$. If for every $z_{0} \in \mathbf{R}$ the sets $A_{z_{0}}$ and $B_{z_{0}}$ have the same area, that is, $\mu\left(A_{z_{0}}\right)=\mu\left(B_{z_{0}}\right)$, then the solids have the same volumes - $\mu(A)=\mu(B)$.

## A. 9 Change of Variables

Remark A.9.1. Throughout this section we'll use always the variable names $D_{t}$ and $D_{x}$ for domains - recall Definition A.2.35.

Definition A.9.2 (Support of a real valued map). Given a map $f: D \rightarrow \mathbf{R}$, where $D \subseteq \mathbf{R}^{n}$ is a domain, the support of $f$, denoted by $\operatorname{supp} f$, is the closure of the collection of points of $D$ with non-zero image under $f$, that is

$$
\operatorname{supp} f:=\operatorname{Cl}(\{x \in D: f(x) \neq 0\})
$$

The goal of this section will be to prove the theorem on the change of variables in the context of multiple integrals. Our approach will be to prove that such theorem holds for elementary $C^{1}$-isomorphisms and also for the compositions of those. We'll decompose the general Riemann integrable map $f: D_{x} \rightarrow \mathbf{R}$ into the composition of such elementary $C^{1}$-isomorphisms and hence prove the theorem. For the proof of the theorem, head to Appendix A.9.

Theorem A.9.3 (Change of variables). Let $D_{x}, D_{t} \subseteq \mathbf{R}^{n}$ be bounded open domains, and $\phi: D_{t} \xrightarrow{\leftrightharpoons} D_{x}$ a $C^{1}$-isomorphism. Let $f: D_{x} \rightarrow \mathbf{R}$ be a Riemann integrable map over $D_{x}$ for which $\operatorname{supp} f$ is compact on $D_{x}$. Then the map given by $t \mapsto f \phi(t)\|\operatorname{det}(\operatorname{Jac} \phi(t))\|$ is Riemann integrable on $D_{t}$ and

$$
\int_{D_{x}} f(x) \mathrm{d} x=\int_{D_{t}} f \phi(t)\|\operatorname{det}(\operatorname{Jac} \phi(t))\| \mathrm{d} t .
$$

## Measurable Sets and Smooth Mappings

Lemma A.9.4. Let $\phi: D_{t} \xrightarrow{\simeq} D_{x}$ be a $C^{1}$-isomorphism. Then the following is true:
(a) Given any subset $E_{t} \subseteq D_{t}$ of Lebesgue measure zero, the image $\phi\left(E_{t}\right) \subseteq D_{x}$ is also a set of Lebesgue measure zero.
(b) Let $E_{t} \subseteq D_{t}$ be any open subset with closure $\mathrm{Cl}_{t} \subseteq D_{t}$ of Jordan measure zero. Then its image $\phi\left(E_{t}\right)$ and the closure $\mathrm{Cl}\left(\phi\left(E_{t}\right)\right)$ are both contained in $D_{x}$ and the closure is a set of Jordan measure zero.
(c) Let $E_{t} \subseteq D_{t}$ be a Jordan-measurable set, and $\mathrm{Cl} E_{t} \subseteq D_{t}$. The image $\phi\left(E_{t}\right)$ is Jordan measurable and $\mathrm{Cl}\left(\phi\left(E_{t}\right)\right) \subseteq D_{x}$.

Proof. (a) Let $I \subseteq D_{t}$ be a closed interval with $E_{t} \subseteq I$. Since $\phi$ is continuously differentiable, there exists $M>0$ such that $\|\mathrm{d} \phi(t)(x)\| \leqslant M$ for all $x \in I$. Using Theorem A.3.30, we obtain that for all pairs $x, y \in I$, we have the relation $\| f(x)$ $f(y)\left\|_{\mathbf{R}^{n}} \leqslant M\right\| x-y \|_{\mathbf{R}^{n}}$. Given any $\varepsilon>0$, since $E_{t}$ is of Lebesgue measure zero, let $\left\{I_{j}\right\}_{j \in J}$ be a countable collection of open intervals covering $E_{t}$ with $\sum_{j \in J} \operatorname{vol} I_{j}<\varepsilon$ and we may assume that $I_{j} \subseteq I$ for all $j \in J$. Consider the countable collection $\left\{\phi\left(I_{j}\right)\right\}_{j \in J}$, which forms a cover for $\phi\left(E_{t}\right)$. Let $t_{j} \in I_{j}$ be the centre point of the interval corresponding to the index $j \in J$ - notice that, since the image of pairs of points of $D_{t}$ under $\phi$ is bounded by the product of their distance and $M$ defining $x_{j}:=\phi\left(t_{j}\right)$, we can consider the open interval $I_{j}^{\prime} \subseteq E_{t}$ with $x_{j}$ being its centre point, with dimension $M$ times of $I_{j}$, then clearly $I_{j}^{\prime}$ is able to cover the whole set $\phi\left(I_{j}\right)$. Let $\left\{I_{j}^{\prime}\right\}_{j}$ be the collection of those intervals we constructed and notice that the image $\phi\left(E_{t}\right)$ is completely encompassed by the union $\bigcup_{j \in J} I_{j}^{\prime}$, that is, $\left\{I_{j}^{\prime}\right\}_{j \in J}$ covers $\phi\left(E_{t}\right)$. Moreover, we the volume of such collection was constructed so that $\sum_{j \in J} \operatorname{vol} I_{j}^{\prime}=\sum_{j \in J} M^{n}$ vol $I_{j}<M^{n} \varepsilon$, that is, $\phi\left(E_{t}\right)$ can be covered by a collection of open intervals whose total volume can be chosen to be arbitrarily small volume thus $\phi\left(E_{t}\right)$ is of Lebesgue measure zero.
(b) Since $\mathrm{Cl} E_{t}$ is a set of Jordan measure zero, it follows it follows that $\mathrm{Cl} E_{t}$ is also of Lebesgue measure zero. Notice that, by the above item, the image $\phi\left(\mathrm{Cl}_{t}\right)$ will also be of Lebesgue measure zero. Notice that $\mathrm{Cl}\left(\phi\left(\mathrm{Cl} E_{t}\right)\right)$ is a compact set of $\mathbf{R}^{n}$ and therefore we can apply Corollary A.7.12 to conclude that $\mathrm{Cl}\left(\phi\left(\mathrm{Cl} E_{t}\right)\right)$ is a set of Jordan measure zero. Since any subset of a set of Jordan measure zero also has measure zero, it follows that $\phi\left(\mathrm{Cl} E_{t}\right)$ is of measure zero.
(c) Since $E_{t}$ is Jordan-measurable, the boundary $\partial E_{t}$ is of Jordan measure zero, therefore so is its image $\phi\left(\partial E_{t}\right)$ by the use of the last item. Since $\phi$ is a $C^{1}$-isomorphism, given any $p \in \operatorname{Int} E_{t}$, we have $\phi(p) \in \operatorname{Int} \phi\left(E_{t}\right)$, therefore $\partial \phi\left(E_{t}\right)=\phi\left(\partial E_{t}\right)$ - thus $\phi\left(E_{t}\right)$ has boundary of Jordan measure zero and therefore is Jordan-measurable.

Corollary A.9.5. In the context of Theorem A.9.3, the integral

$$
\int_{D_{t}} f \phi(t)\|\operatorname{det}(\operatorname{Jac} \phi(t))\| \mathrm{d} t
$$

exists.
Proof. Since $\phi$ is a $C^{1}$-isomorphism, its Jacobian Jac $\phi(t)$ is invertible for every $t \in D_{t}$, thus $\operatorname{det}(\operatorname{Jac} \phi(t)) \neq 0$. If $g: D_{t} \rightarrow \mathbf{R}$ is defined as $g(t):=f \phi(t)\|\operatorname{det}(\operatorname{Jac} \phi(t))\|$, then for every $t \in D_{t}$ such that $\phi(t) \notin \operatorname{supp} f$ we have $t \notin \operatorname{supp} g-$ thus supp $g=$ $\operatorname{supp}\left(f \phi \phi^{-1}(\operatorname{supp} f)\right)$, and $\operatorname{supp} g$ is a compact subset of $D_{t}$ since it is closed. From this we can conclude that the map $g_{\chi_{D_{t}}}$ has points of discontinuity in the fibers $\phi^{-1}(x)$, where $x \in D_{x}$ is a point of discontinuity of $f$. Since $f$ is Riemann integrable in $D_{x}$, such points form a Lebesgue measure zero set - thus the points of discontinuity of $g_{\chi_{D_{t}}}$ are also of Lebesgue measure zero since $\phi$ preserves Lebesgue measure zero sets by Lemma A.9.4. Using the Lebesgue's criterion, we conclude that $g_{\chi_{D_{t}}}$ is Riemann integrable over any closed interval $I_{t}$ such that $D_{t} \subseteq I_{t}$ and thus the said integral does exist.

## One-Dimensional Change of Variables

Lemma A.9.6 (Changing variables in one dimension). Let $\phi: I_{t} \xrightarrow{\simeq} I_{x}$ be a $C^{1}$-isomorphism over the 1-dimensional intervals $I_{t}, I_{x} \subseteq \mathbf{R}$ and let $f: I_{x} \rightarrow \mathbf{R}$ be a Riemann integrable map over $I_{x}$. Then the map $I_{t} \ni t \mapsto f \phi(t)\left\|\phi^{\prime}(t)\right\|$ is a Riemann integrable map over $I_{t}$ and the following relation holds

$$
\int_{I_{x}} f(x) \mathrm{d} x=\int_{I_{t}} f \phi(t)\left\|\phi^{\prime}(t)\right\| \mathrm{d} t
$$

Proof. All maps $f, \phi$ and $\phi^{\prime}$ are bounded, thus the map $g: I_{t} \rightarrow \mathbf{R}$ given by $g(t):=$ $f \phi(t)\left\|\phi^{\prime}(t)\right\|$ is bounded. As pointed out in the proof of Corollary A.9.5, the points of discontinuity of $g$ are only those lying in the fibers $\phi^{-1}(x)$, where $x \in I_{x}$ is a point of discontinuity of $f$ - since $f$ is Riemann integrable and $\phi^{-1}$ is a $C^{1}$-isomorphism, the points of discontinuity of $g$ form a set of Lebesgue measure zero, thus $g$ is Riemann integrable over $I_{t}$. Let $P_{x}:=\left\{\left[x_{j-1}, x_{j}\right]\right\}_{j \in J}$ be a partition of $I_{x}$ and define $P_{t}:=\left\{\phi^{-1}\left(\left[x_{j-1}, x_{j}\right]\right)\right\}_{j \in J}$ as the partition of $I_{t}$. Since $\phi$ and $\phi^{-1}$ are both continuous functions over compact sets, it follows that they are both uniformly continuous thus, taking a sequence of partitions of $I_{x}$ and the induced partitions of $I_{t}$, we have mesh $P_{x} \rightarrow 0$ if and only if mesh $P_{t} \rightarrow 0$. Let $\xi$ be a collection of distinguished points
of $P_{x}$ and $\tau:=\left\{\phi^{-1}\left(\xi_{j}\right)\right\}_{j \in J}$ be the corresponding collection of distinguishes points of $P_{t}$. Notice that, from construction, we have the following

$$
\sum_{j \in J} f\left(\xi_{j}\right)\left\|x_{j}-x_{j-1}\right\|=\sum_{j \in J} f \phi\left(\tau_{j}\right)\left\|\phi\left(t_{j}\right)-\phi\left(t_{j-1}\right)\right\|
$$

Since both $f$ and $g$ are Riemann integrable (over their respective domains), we find that the choice of distinguished points is arbitrary, thus we may choose $\xi$ so that, when applying the mean value theorem to $\phi$, we have $\phi^{\prime}\left(\tau_{j}\right)=\frac{\phi\left(t_{j}\right)-\phi\left(t_{j-1}\right)}{t_{j}-t_{j-1}}$ - where, as before, $\tau_{j}:=\phi^{-1}\left(\xi_{j}\right)$. This way we may rewrite the equation as

$$
\sum_{j \in J} f\left(\xi_{j}\right)\left\|x_{j}-x_{j-1}\right\|=\sum_{j \in J} f \phi\left(\tau_{j}\right)\left\|\phi^{\prime}\left(\tau_{j}\right)\right\|\left\|t_{j}-t_{j-1}\right\|
$$

Now, taking the limit mesh $P_{x} \rightarrow 0$, which implies mesh $P_{t} \rightarrow 0$, we find the integral equality just wanted.
Corollary A.9.7. Let $\phi: I_{t} \xrightarrow{\simeq} I_{x}$ be a $C^{1}$-isomorphism between closed 1-dimensional intervals $I_{t}, I_{x} \subseteq \mathbf{R}$, and $f: I_{x} \rightarrow \mathbf{R}$ be a Riemann integrable map over $I_{x}$. Then we have the following relations

$$
\overline{\int_{I_{x}}} f(x) \mathrm{d} x=\overline{\int_{I_{t}}} f \phi(t)\left\|\phi^{\prime}(t)\right\| \mathrm{d} t \text { and } \underline{\int_{I_{x}}} f(x) \mathrm{d} x=\underline{\int_{I_{t}}} f \phi(t)\left\|\phi^{\prime}(t)\right\| \mathrm{d} t
$$

Proof. We'll verify the first relation, the second may be obtained with an analogous proof. Construct partitions $P_{x}$ and $P_{t}$ for $I_{x}$ and $I_{t}$, respectively, just as we've done for the proof of Lemma A.9.6. Assume for the time being that $f$ is a non-negative map (we'll take hold of the general case just in a bit) and let $M>0$ be a constant bounding $f$. Define $\varepsilon:=\sup _{j \in J} \omega\left(\phi,\left[t_{j-1}, t_{j}\right]\right)$, that is, the supremum oscillation of $\phi$ over each of the intervals of the partition $P_{t}$. Notice that we have

$$
\begin{aligned}
\sum_{j \in J} \sup _{x \in\left[x_{j-1}, x_{j}\right]} f(x)\left\|x_{j}-x_{j-1}\right\| & \leqslant \sum_{j \in J} \sup _{t \in\left[t_{j-1}, t_{j}\right]} f \phi(t) \sup _{t \in\left[t_{j-1}, t_{j}\right]}\left\|\phi^{\prime}(t)\right\|\left\|t_{j}-t_{j-1}\right\| \\
& \leqslant \sum_{j \in J} \sup _{t \in\left[t_{j-1}, t_{j}\right]}\left[f \phi(t) \sup _{t \in\left[t_{j-1}, t_{j}\right]}\left\|\phi^{\prime}(t)\right\|\right]\left\|t_{j}-t_{j-1}\right\| \\
& \leqslant \sum_{j \in J} \sup _{t \in\left[t_{j-1}, t_{j}\right]} f \phi(t)\left(\left\|\phi^{\prime}(t)\right\|+\varepsilon\right)\left\|t_{j}-t_{j-1}\right\| \\
& \leqslant \sum_{j \in J} \sup _{t \in\left[t_{j-1}, t_{j}\right]} f \phi(t)\left\|\phi^{\prime}(t)\right\|\left\|t_{j}-t_{j-1}\right\|+\varepsilon \sum_{j \in J} \sup _{t \in\left[t_{j-1}, t_{j}\right]} f \phi(t)\left\|t_{j}-t_{j-1}\right\| \\
& \leqslant \sum_{j \in J} \sup _{t \in\left[t_{j-1}, t_{j}\right]} f \phi(t)\left\|\phi^{\prime}(t)\right\|\left\|t_{j}-t_{j-1}\right\|+\varepsilon M \operatorname{vol} I_{t}
\end{aligned}
$$

where vol $I_{t}$ happens to be the length of the interval, since $I_{t} \subseteq \mathbf{R}$. Since $\phi$ is uniformly continuous, we find, from the last inequalities, that in the limit mesh $P_{t} \rightarrow 0$ (which implies in mesh $P_{x} \rightarrow 0$ ):

$$
\bar{\int}_{I_{x}} f(x) \mathrm{d} x \leqslant \bar{\int}_{I_{t}} f \phi(t)\left\|\phi^{\prime}(t)\right\| \mathrm{d} t
$$

Notice that we can use the same analogous proof for the $C^{1}$-isomorphism $\phi^{-1}$ and the map $I_{t} \ni t \mapsto f \phi(t)\left\|\phi^{\prime}(t)\right\|$, we find that

$$
\int_{I_{t}} f \phi(t)\left\|\phi^{\prime}(t)\right\| \mathrm{d} t \leqslant \overline{\int_{I_{x}}} f(x) \mathrm{d} x
$$

thus the equality for the first relation has been established for non-negative $f$. Notice, however, that $f=\max (f, 0)-\max (-f, 0)$, for any $\operatorname{map} f$, thus the equality holds for the general case.

## Change of Variables on Elementary $C^{1}$-Isomorphisms

Definition A. 9.8 (Elementary $C^{1}$-isomorphism in $\mathbf{R}^{n}$ ). Let $\left\{t_{j}\right\}_{j=1}^{n}$ and $\left\{x_{j}\right\}_{j=1}^{n}$ be basis for the euclidean space $\mathbf{R}^{n}$. Given any $1 \leqslant k \leqslant n$, a $C^{1}$-isomorphism $\phi: D_{t} \xrightarrow{\simeq} D_{x}$ such that $\phi_{j}\left(t_{j}\right)=t_{j}$ for every $j \neq k$, and $\phi_{k}\left(t_{k}\right)=x_{k}$, is said to be a $k$-elementary $C^{1}$-isomorphism - that is, the only coordinate of $\mathbf{R}^{n}$ changed under the mapping of $\phi$ is the $k$-th.

Lemma A.9.9. In the context of Theorem A.9.3, if $\phi$ is an elementary $C^{1}$-isomorphism, then the proposition is valid.

Proof. Let $\phi$ be an elementary $C^{1}$-isomorphism on the $k$-th coordinate. Given points $x, t \in \mathbf{R}^{n}$, we define $x^{\prime}:=\left(x_{j}\right)_{j=1, j \neq k}^{n} \in \mathbf{R}^{n-1}$ and $t^{\prime}:=\left(t_{j}\right)_{j=1, j \neq k}^{n} \in \mathbf{R}^{n-1}$. Moreover, for every $x^{\prime}, t^{\prime} \in \mathbf{R}^{n-1}$, we define 1-dimensional sections of the domains $D_{x}$ and $D_{t}$, respectively, as

$$
\begin{aligned}
D_{x_{k}}\left(x^{\prime}\right) & :=\left\{p \in D_{x}: p_{k}=x_{k} \text { and } p_{j}=x_{j}^{\prime} \text { for } j \neq k\right\}, \\
D_{t_{k}}\left(t^{\prime}\right) & :=\left\{p \in D_{t}: p_{k}=t_{k} \text { and } t_{j}=t_{j}^{\prime} \text { for } j \neq k\right\} .
\end{aligned}
$$

Let $I_{x} \subseteq \mathbf{R}^{n}$ be a closed interval containing $D_{x}$ and we let $I_{x^{\prime}}^{0} \times I_{x_{k}} \times I_{x^{\prime}}^{1}:=I_{x}$ be the representation of $I_{x}$ by closed intervals $I_{x^{\prime}}^{0} \subseteq \mathbf{R}^{k-1}, I_{x_{k}} \subseteq \mathbf{R}$ and $I_{x^{\prime}}^{1} \subseteq \mathbf{R}^{n-k}$. Analogously, let $I_{t} \subseteq \mathbf{R}^{n}$ be a closed interval containing $D_{t}$ and let $I_{t^{\prime}}^{0} \times I_{t_{k}} \times I_{t^{\prime}}^{1}:=I_{t}$ where, as before,
$I_{t^{\prime}}^{0} \subseteq \mathbf{R}^{k-1}, I_{t_{k}} \subseteq \mathbf{R}$ and $I_{t^{\prime}}^{1} \subseteq \mathbf{R}^{n-k}$ are all closed intervals. Thus, we have

$$
\begin{align*}
\int_{D_{x}} f(x) \mathrm{d} x & =\int_{I_{x}} f_{\chi_{D_{x}}}(x) \mathrm{d} x \\
& =\int_{I_{x^{\prime}}^{0} \times I_{x^{\prime}}^{1}}\left(\int_{I_{x_{k}}} f_{\chi_{D_{x}}}\left(x^{\prime}\right) \mathrm{d} x_{k}\right) \mathrm{d} x^{\prime}  \tag{A.46}\\
& =\int_{I_{x^{\prime}}^{0} \times I_{x^{\prime}}^{1}}\left(\int_{D_{x_{k}}\left(x^{\prime}\right)} f(x) \mathrm{d} x\right) \mathrm{d} x_{0}^{\prime}  \tag{A.47}\\
& =\int_{I_{t^{\prime}}^{0} \times I_{t^{\prime}}^{1}}\left(\int_{I_{t_{k}}} f \phi_{\chi_{D_{t}}}\left(t_{1^{\prime}}^{\prime}, \ldots, t_{k}, \ldots, t_{n}^{\prime}\right)\left\|\operatorname{det}\left(\operatorname{Jac} \phi_{\chi_{D_{t}}}\left(t_{1^{\prime}}^{\prime}, \ldots, t_{k}, \ldots, t_{n}^{\prime}\right)\right)\right\| \mathrm{d} t_{k}\right) \mathrm{d} t^{\prime} \\
& =\int_{I_{t^{\prime}}^{0} \times I_{I_{k}} \times \times I_{t^{\prime}}^{1}} f \phi_{\chi_{D_{D_{t}}}}(t)\left\|\operatorname{det}\left(\operatorname{Jac} \phi_{\chi_{D_{t}}}(t)\right)\right\| \mathrm{d} t=\int_{I_{t}} f \phi_{\chi_{D_{t}}}(t)\left\|\operatorname{det}\left(\operatorname{Jac} \phi_{\chi_{D_{t}}}(t)\right)\right\| \mathrm{d} t  \tag{A.48}\\
& =\int_{D_{t}} f \phi(t)\|\operatorname{det}(\operatorname{Jac} \phi(t))\| \mathrm{d} t \tag{A.50}
\end{align*}
$$

where we have the following use of theorems and definitions for each of the equations: Eq. (A.46) used Theorem A.8.4, Eq. (A.47) used Definition A.7.20, Eq. (A.48) used Lemma A.9.6, Eq. (A.49) used Theorem A.8.4, and, finally, Eq. (A.50) used Definition A.7.20. This proves the lemma.

Proposition A.9.10. Let $f: E \rightarrow \mathbf{R}^{n}$ be a $C^{1}$ map on the open set $E \subseteq \mathbf{R}^{n}$, with $0 \in$ $E$ - moreover, we impose that $f(0):=0$ and $\mathrm{d} f(0): T_{0} \mathbf{R}^{n} \rightarrow T_{0} \mathbf{R}^{n}$ is an $\mathbf{R}$-linear isomorphism. Then there exists a neighbourhood $U \subseteq E$ of $0 \in \mathbf{R}^{n}$, with mappings:

- $\left\{\phi_{j}: U \rightarrow \mathbf{R}^{n}\right\}_{j=1}^{n}$ of elementary $C^{1}$-isomorphisms, such that $\phi_{j}(0)=0$ and $\mathrm{d} \phi(0)$ is an $\mathbf{R}$-linear isomorphism.
- We define, for each $1 \leqslant k \leqslant n$, the map $\tau_{k}: \mathbf{R}^{n} \rightarrow \mathbf{R}^{n}$ as $\tau_{k}\left(\sum_{j=1}^{n} x_{j} e_{j}\right):=$ $\sum_{j=1}^{n} x_{\tau(j)} e_{j}$, where $\tau \in \operatorname{Aut}_{G r \mathrm{p}}([n])$ is a transposition.

These maps allow us to write $\left.f\right|_{U}$ as the composition of elementary $C^{1}$-isomorphisms and transpositions

$$
f(x)=\tau_{1} \ldots \tau_{n-1} g_{n} \ldots g_{1}(x), \text { for all } x \in U
$$

Proof. Define, for every $0 \leqslant k \leqslant n$, the map $p_{k}: \mathbf{R}^{n} \rightarrow \mathbf{R}^{n}$ by $p_{k}\left(\sum_{j=1}^{n} x_{j} e_{j}\right):=\sum_{j=1}^{k} x_{j} e_{j}$.
We do induction in the following proposition, where $1 \leqslant m \leqslant n-1$

- There exists a neighbourhood $V_{m} \subseteq \mathbf{R}^{n}$ of zero, a map $f_{m} \in C^{1}\left(V_{m}\right)$ - with $f_{m}(0)=0$ and $\mathrm{d} f_{m}(0)$ linear isomorphism (we define $f_{1}:=f$ )- for which

$$
p_{m-1} f_{m}(x)=p_{m-1}(x), \text { for all } x \in V_{m} .
$$

Notice that the proposition is clearly true for $m=1$, now, we assume it's true for $1<m<n-1$ as our hypothesis of induction.

Notice that $f_{m}$ is a map that does only change the last $x_{m}, \ldots, x_{n}$ variables of its input, which allows us to write it as

$$
f_{m}(x)=p_{m-1}(x)+\sum_{j=m}^{n} \alpha_{j}(x) e_{j}
$$

where $\alpha_{j}: V_{m} \rightarrow \mathbf{R}$ is a $C^{1}\left(V_{m}\right)$ map for all $m \leqslant j \leqslant n$, moreover, this implies in

$$
\mathrm{d} f_{m}(0)\left(e_{m}\right)=\sum_{j=m}^{n} \partial_{m} \alpha_{j}(0) e_{j} .
$$

Since $\mathrm{d} f_{m}(0)$ is an isomorphism, it cannot be the zero-map, thus there must exist some $m \leqslant k_{0} \leqslant n$ for which $\partial_{m} \alpha_{k_{0}}(0) \neq 0$. Define $\tau_{m}$ to be the transposition interchanging the $m$-th and $k_{0}$-th values of its input. Define a map $\phi_{m}: V_{m} \rightarrow \mathbf{R}^{n}$ as

$$
\phi_{m}(x):=x+\left(\alpha_{k_{0}}(x)-x_{m}\right) e_{m}
$$

so that $\phi_{m} \in C^{1}\left(V_{m}\right)$, and only changes the $m$-th coordinate of its input. Moreover $\mathrm{d} \phi_{m}(0)$ is an isomorphism, thus there must exist a neighbourhood $U_{m} \subseteq V_{m}$ of zero for which the induced $\operatorname{map} \phi_{m}: U_{m} \rightarrow \phi_{m}\left(U_{m}\right)$ is a $C^{1}$-isomorphism (which, in fact, is elementary). We define $V_{m+1}:=\phi_{m}\left(U_{m}\right)$, and also

$$
f_{m+1}(y):=\tau_{m} f_{m} \phi_{m}^{-1}(y), \text { for all } y \in V_{m+1}
$$

which makes $f_{m+1} \in C^{1}\left(V_{m+1}\right)$, also $f_{m+1}(0)=0$, and $\mathrm{d} f_{m+1}(0)$ is an isomorphism. Notice that, for all $x \in U_{m}$ we have

$$
\begin{aligned}
p_{m} f_{m+1}\left(\phi_{m}(x)\right) & =p_{m} \tau_{m} f_{m} \phi_{m}^{-1}\left(\phi_{m}(x)\right) \\
& =p_{m} \tau_{m} f(x) \\
& =p_{m} \tau_{m}\left[p_{m-1}(x)+\sum_{j=m}^{n} \alpha_{j}(x) e_{j}\right] \\
& =p_{m}\left[p_{m-1}(x)+\alpha_{k_{0}}(x) e_{m}+\cdots+\alpha_{m}(x) e_{k_{0}}+\cdots+\alpha_{n}(x) e_{n}\right] \\
& =p_{m-1}(x)+\alpha_{k_{0}}(x) e_{m} \\
& =p_{m} \phi_{m}(x) .
\end{aligned}
$$

That is, since $\phi_{m}$ is a bijection in $V_{m+1}$, for any $y \in V_{m+1}$ we have the equality

$$
p_{m} f_{m+1}(y)=p_{m}(y)
$$

This concludes the proof by induction of our initial statement. We now must show that it, in fact, implies the proposition we set out to prove. For that we recall the definition of $f_{m+1}$ and notice that, by taking $\tau_{m}$ from both sides we obtain $\tau_{m} f_{m+1}(y)=$
$\tau_{m} \tau_{m} f_{m} \phi_{m}^{-1}(y)=f_{m} \phi_{m}^{-1}(y)$ for every $y \in V_{m+1}$ - which allow us to write, taking $y=\phi_{m}(x)$,

$$
f_{m}(x)=\tau_{m} f_{m+1}\left(\phi_{m}(x)\right), \text { for every } x \in U_{m}
$$

This recursive definition implies in (recalling that we set $f_{1}:=f$ ):

$$
f=\tau_{1} f_{2} \phi_{1}=\cdots=\tau_{1} \ldots \tau_{n-1} f_{n} \phi_{n-1} \ldots \phi_{1}
$$

which is what we wanted since $f_{n}$ is an elementary $C^{1}$-isomorphism for some neighbourhood of zero.

## Proof for the Theorem on Change of Variables

We now go for our (finally) last lemma before we can get to the proof of the main theorem of the section.

Lemma A.9.11. Let $D_{s} \xrightarrow{\psi} D_{t} \xrightarrow{\phi} D_{x}$ be $C^{1}$-isomorphisms. Moreover, let $f: D_{x} \rightarrow \mathbf{R}$ be a Riemann integrable map over $D_{x}$. If Theorem A.9.3 holds for both $\psi$, and for $\phi$, then it is also valid for the composition $\phi \psi: D_{s} \xrightarrow{\simeq} D_{x}$.

Proof. Notice that, for every $s \in D_{s}, \operatorname{Jac}(\phi \psi)(s)=\operatorname{Jac} \phi(\psi(s)) \operatorname{Jac} \psi(s)$, then

$$
\operatorname{det}(\operatorname{Jac}(\phi \psi)(s))=\operatorname{det}(\operatorname{Jac} \phi(\psi(s))) \operatorname{det}(\operatorname{Jac} \psi(s))
$$

Therefore we can write

$$
\begin{aligned}
\int_{D_{x}} f(x) \mathrm{d} x=\int_{D_{t}} f \phi(t)\|\operatorname{det}(\operatorname{Jac} \phi(t))\| \mathrm{d} t & =\int_{D_{s}} f \phi \psi(s)\|\operatorname{det}(\operatorname{Jac} \phi(\psi(s)))\|\|\operatorname{det}(\operatorname{Jac} \psi(s))\| \mathrm{d} s \\
& =\int_{D_{s}} f \phi \psi(s)\|\operatorname{det}(\operatorname{Jac}(\phi \psi)(s))\| \mathrm{d} s,
\end{aligned}
$$

which proves the proposition.
Now we are ready for the proof of Theorem A.9.3.
Proof. Define the compact set $K_{t}:=\operatorname{supp}(t \mapsto f \phi(t)\|\operatorname{det}(\operatorname{Jac} \phi(t))\|)$, and, for every $t \in K_{t}$, let $U(t) \subseteq D_{t}$ be a neighbourhood of $t$ with diameter less than $\delta(t)$, for some $\delta(t)>0$ - for which $\left.\phi\right|_{U(t)}$ can be decomposed into elementary $C^{1}$-isomorphisms. For each $t \in K_{t}$, let $U^{\prime}(t) \subseteq U(t)$ be a neighbourhood of $t$ with diameter less than $\delta(t) / 2$ - with this, the collection $\left\{U^{\prime}(t)\right\}_{t \in K_{t}}$ is a cover for $K_{t}$, more than that, since $K_{t}$ is compact, there exists a finite collection of points $\left\{t_{j}\right\}_{j=1}^{k}$ such that $\left\{U^{\prime}\left(t_{j}\right)\right\}_{j=1}^{k}$ is a cover for $K_{t}$. Define $\delta:=\frac{1}{2} \min \left(\delta\left(t_{j}\right)_{j=1}^{k}\right)$ so that, for any set $A \subseteq D_{t}$ with closure of diameter less than $\delta$ and non-empty intersection with the support, $A \cap K_{t} \neq \varnothing$, the given set must be contained in some $U^{\prime}\left(t_{m}\right)$ for some $1 \leqslant m \leqslant k$.

Let $I \subseteq \mathbf{R}^{n}$ be an interval containing $D_{t}$, and $P:=\left\{I_{s}\right\}_{s \in S}$ a partition of $I$ with mesh $P<\min (\delta, d)$ - where $d:=\inf _{(a, b) \in \partial K_{t} \times \partial D_{t}\|b-a\| \text { is the minimal distance between }}$ the boundaries of $K_{t}$ and $D_{t}$. Let $S^{\prime}$ be the index set such that $I_{s} \cap K_{t} \neq \varnothing$ if and only
if $s \in S^{\prime}$ - the set indices of intervals of the partition $P$ with non-empty intersection with the support. Then

$$
\begin{aligned}
\int_{D_{t}} f \phi(t)\|\operatorname{det}(\operatorname{Jac} \phi(t))\| \mathrm{d} t & =\int_{I} f \phi_{\chi_{D_{t}}}(t)\left\|\operatorname{det}\left(\operatorname{Jac} \phi_{\chi_{D_{t}}}(t)\right)\right\| \mathrm{d} t \\
& =\sum_{s^{\prime} \in S^{\prime}} \int_{I_{s^{\prime}}} f \phi(t)\|\operatorname{det}(\operatorname{Jac} \phi(t))\| \mathrm{d} t .
\end{aligned}
$$

Since $I_{s} \in P$ are Jordan-measurable sets, so is their image $\phi\left(I_{s}\right) \subseteq D_{x}$ - this follows from Lemma A.9.4. Define the Jordan-measurable set $E:=\bigcup_{s^{\prime} \in S^{\prime}} I_{s^{\prime}}$ so that, by our construction, given any $x \in \operatorname{supp} f$, let $t:=\phi^{-1}(x)$, then $f \phi(t)\|\operatorname{det}(\operatorname{Jac} \phi(t))\| \neq 0$ thus $t \in K_{t}$ - therefore, $t \in \bigcup_{s^{\prime} \in \mathcal{S}^{\prime}} I_{s^{\prime}}$, which implies in $x \in E$, that is, $\operatorname{supp} f \subseteq E$. Notice that, if $I_{x} \subseteq \mathbf{R}^{n}$ is an interval containing $D_{x}$, then

$$
\begin{align*}
\int_{D_{x}} f(x) \mathrm{d} x & =\int_{I_{x}} f_{\chi_{D_{x}}}(x) \mathrm{d} x=\underbrace{\int_{I_{x} \backslash E} f_{\chi_{D_{x}}}(x) \mathrm{d} x}+\int_{E} f_{\chi_{D_{x}}}(x) \mathrm{d} x \\
& =\int_{E} f(x) \mathrm{d} x=\sum_{s^{\prime} \in S^{\prime}} \int_{\phi\left(I_{s^{\prime}}\right)} f(x) \mathrm{d} x . \tag{A.51}
\end{align*}
$$

Since $\phi$ is decomposed into elementary $C^{1}$-isomorphisms in any neighbourhood $U^{\prime}(t)$, for $t \in K_{t}$, it follows that $\phi$ decomposes into elementary $C^{1}$-isomorphisms in every interval $I_{s^{\prime}}$ for $s^{\prime} \in S^{\prime}$ - hence, by means of Lemma A.9.9 we conclude that, for every $s^{\prime} \in S^{\prime}$

$$
\begin{equation*}
\int_{\phi\left(I_{s^{\prime}}\right)} f(x) \mathrm{d} x=\int_{I_{s^{\prime}}} f \phi(t)\|\operatorname{det}(\operatorname{Jac} \phi(t))\| \mathrm{d} t \tag{A.52}
\end{equation*}
$$

Thus, merging both Eq. (A.51) and Eq. (A.52), we obtain the desired equality.

## A. 10 An Elementary Construction of Differential Forms

Definition A.10.1 ( $k$-surface). Let $E \subseteq \mathbf{R}^{n}$ be an open set. We define a $k$-surface in $E$ to be a $C^{1} \operatorname{map} \Phi: D \rightarrow E$, where $D \subseteq \mathbf{R}^{k}$ is a compact set - this set is commonly referenced to as the parameter domain of $\Phi$.
Definition A.10.2 (Differential $k$-form). Let $E \subseteq \mathbf{R}^{n}$ be an open set. A differential $k$-form in $E$, for $k>0$, is a multilinear map $\omega$ given by

$$
\omega:=\sum f_{j_{1} \ldots j_{k}} \mathrm{~d} x_{j_{1}} \wedge \cdots \wedge \mathrm{~d} x_{j_{k}}
$$

where the sum runs over the sequence of indices $\left(j_{1}, \ldots, j_{k}\right)$, where $1 \leqslant j_{r} \leqslant n$ for each $1 \leqslant r \leqslant k$, and $f_{j_{1} \ldots j_{k}}: E \rightarrow \mathbf{R}$ are continuous maps - that is, 0 -forms. A differential $k$-form is said to be of class $C^{p}$ if each map $f_{j_{1} \ldots j_{k}}$ is of class $C^{p}$.

The form $\omega$ is said to be a basic $k$-form if we have the ordering $1 \leqslant j_{1}<\cdots<j_{k} \leqslant n$ for its indices - in this case we commonly denote each sequence of indexes as $I:=$ $\left(j_{1}, \ldots, j_{k}\right)$ and write

$$
\omega=\sum_{I} f_{I} \mathrm{~d} x_{I}
$$

Definition A.10.3 (Line integral). The integrals of 1-forms are called line integrals.
Example A.10.4. Lets calculate a line integral on the 1 -surface ( $C^{1}$ curve) $\gamma:[0,2 \pi] \rightarrow$ $\mathbf{R}^{2}$ defined by $t \mapsto(a \cos (t), b \sin (t))$, for constants $a, b>0$. Let $\omega=x \mathrm{~d} y$ be a 1 -form on $\mathbf{R}^{2}$. Measuring this curve with respect to $\omega$ yields

$$
\int_{\gamma} \omega=\int_{0}^{2 \pi} a \cos (t)[b \cos (t)] \mathrm{d} t=a b \int_{0}^{2 \pi} \frac{\cos (2 t)-1}{2} \mathrm{~d} t=-a b \pi
$$

Definition A.10.5 (Surface area). Let $\Phi: D \rightarrow E$ be a $k$-surface on the open set $E \subseteq \mathbf{R}^{n}$, and $\omega:=\sum_{I} f_{i_{1} \ldots i_{k}} \mathrm{~d} x_{i_{1} \ldots i_{k}}$. We define the area of the surface $\Phi$ on $\omega$ to be given by

$$
\int_{\Phi} \omega:=\int_{D} \sum_{I} f_{i_{1} \ldots i_{k}}(\Phi(t)) \operatorname{det}\left[\partial_{j} \Phi_{i_{r}}(t)\right]_{1 \leqslant r, j \leqslant k} \mathrm{~d} t
$$

Definition A.10.6 (Piecewise smooth). We define the concept of a piecewise smooth surface in an inductive manner. A point is a zero dimensional surface of any smoothness class. A surface $S \subseteq \mathbf{R}^{n}$ of dimension $k$ is piecewise smooth if, after a countable collection of at most $(k-1)$-dimensional piecewise smooth surfaces can be removed from $S$, the resulting surface can be decomposed into a countable collection of $k$ dimensional smooth surfaces.

Theorem A.10.7 (A zero form has null coefficients). Let $\omega:=\sum_{J} f_{J} \mathrm{~d} x_{J}$ be a $k$-form in an open set $E \subseteq \mathbf{R}^{n}$, in its standard representation. If $\omega=0$ in $E$ (that is $\omega(\Phi)=0$ for all $k$-surface $\Phi$ in $E$ ), then $f_{J}(x)=0$ for all $x \in E$.

Proof. Suppose, for the sake of contradiction, that there exists an index sequence $S:=$ $\left(s_{1}, \ldots, s_{k}\right)$ such that $f_{S}(y) \neq 0$ for some $y \in E$. Since $f_{S}$ is continuous, let $\delta>0$ be such that $f_{J}(x)>0$ whenever $\left\|x_{i}-y_{i}\right\| \leqslant \delta$, for all $1 \leqslant i \leqslant n$. Let $D \subseteq \mathbf{R}^{k}$ be the compact set given by

$$
D:=\left\{t \in \mathbf{R}^{k}:\left\|t_{j}\right\| \leqslant \delta \text { for all } 1 \leqslant j \leqslant k\right\}
$$

Define a $k$-surface $\Phi: D \rightarrow E$ to be given by

$$
\Phi(t):=v+\sum_{j=1}^{k} t_{j} e_{s_{j}}, \text { for all } t \in D
$$

Notice that $D$ was chosen so that $f_{S}(\Phi(t))>0$ for all $t \in D$. Notice that

$$
\int_{\Phi} \omega=\int_{D} \sum_{J} f_{J}(\Phi(t)) \operatorname{det}\left[\partial_{i} \Phi_{j_{r}}(t)\right]_{1 \leqslant i, r \leqslant k} \mathrm{~d} t=\int_{D} f_{S}(\Phi(t)) \operatorname{det}\left[\partial_{i} \Phi_{s_{r}}\right]_{1 \leqslant i, r \leqslant k} \mathrm{~d} t
$$

since for all indexing sequence $J:=\left(j_{1}, \ldots, j_{k}\right) \neq S$ the matrix $\left[\partial_{i} \Phi_{j_{r}}(t)\right]_{1 \leqslant i, r \leqslant k}$ has at least one column equal to zero - since there must exists at least one $j_{r_{0}} \in J \backslash S$, thus $\Phi_{j_{0}}(t)=y_{j_{0}}$ for all $t \in D$ and thus the column $\left[\partial_{i} \Phi_{j_{0}}(t)\right]_{1 \leqslant i \leqslant k}=0$, which implies in
$\operatorname{det}\left[\partial_{i} \Phi_{j_{r}}(t)\right]_{1 \leqslant i, r \leqslant k}=0$. On the other hand, notice that $\partial_{i} \Phi_{s_{r}}(t)=\delta_{i r}$ for all $1 \leqslant i, r \leqslant k$, thus $\operatorname{det}\left[\partial_{i} \Phi_{s_{r}}(t)\right]_{1 \leqslant i, r \leqslant k}=1$. We conclude that

$$
\int_{\Phi} \omega=\int_{D} f_{S}(t) \mathrm{d} t
$$

which is strictly positive since $f_{S}(y)>0$. Therefore $\omega(\Phi) \neq 0$, which is a contradiction - thus there must exist no indexing sequence $S$ and therefore every coefficient is the zero-map.

## Differential Operator

Definition A.10.8. Let $f: E \rightarrow \mathbf{R}$ be a map of class $C^{1}$, we define an operator $d$ which transforms any 0 -form $f$ into

$$
\mathrm{d} f:=\sum_{j=1}^{n} \partial_{j} f \mathrm{~d} x_{j}
$$

Now, for any $k$-form $\omega:=\sum_{J} f_{J} \mathrm{~d} x_{J}$, where $k \geqslant 1$ and $f_{J}: E \rightarrow \mathbf{R}$ is again a map of class $C^{1}$ (a 0 -form) we associate the $(k+1)$-form $\mathrm{d} \omega$ - which is defined by

$$
\mathrm{d} \omega:=\sum_{J} \mathrm{~d} f_{J} \wedge \mathrm{~d} x_{J}
$$

Example A.10.9. Let $E \subseteq \mathbf{R}^{n}$ be an open set, and $\gamma:[0,1] \rightarrow E$ be a 1-surface (that is, a continuous differentiable curve). If we let $f: E \rightarrow \mathbf{R}$ be a $C^{1}$ map, we have from the definition that the integral over the curve $\gamma$ of $\mathrm{d} f$ is given by - recalling Theorem A.3.21,

$$
\int_{\gamma} \mathrm{d} f=\int_{0}^{1} \sum_{j=1}^{n} \partial_{j} f(\gamma(t)) \gamma_{j}^{\prime}(t) \mathrm{d} t=\int_{0}^{1} \mathrm{~d}(f \gamma)(t) \mathrm{d} t=f \gamma(1)-f \gamma(0)
$$

since $f \gamma:[0,1] \rightarrow \mathbf{R}$ is a continuously differentiable map.
Theorem A.10.10. The following are properties of the differential operator on forms. Let $E \subseteq \mathbf{R}^{n}$ be some open set.
(a) (Skew-product rule) Let $\omega$ be a $k$-form and $\gamma$ be a $m$-form, both of class $C^{1}(E)$. Then

$$
\mathrm{d}(\omega \wedge \gamma)=(\mathrm{d} \omega) \wedge \gamma+(-1)^{k} \omega \wedge \mathrm{~d} \gamma
$$

(b) If $\omega$ is a $k$-form of class $C^{2}(E)$, then $\mathrm{d}(\mathrm{d} \omega)=0$.

Proof. (a) Let $\omega:=\sum_{I} f_{I} \mathrm{~d} x_{I}$ and $\gamma(x):=\sum_{J} g_{J} \mathrm{~d} x_{J}$ for $C^{1}$ coefficients $f_{I}, g_{J}: E \rightrightarrows \mathbf{R}$ — if $k$ or $m$ are zero, we just omit the 1 -forms from the definitions. From the wedge product we have

$$
\mathrm{d}(\omega \wedge \gamma)=\sum_{I, J} \mathrm{~d}\left(f_{I} \cdot g_{J} \mathrm{~d} x_{I} \wedge \mathrm{~d} x_{J}\right)=\sum_{I, J} \mathrm{~d}\left(f_{I} \cdot g_{J}\right) \wedge \mathrm{d} x_{I} \wedge \mathrm{~d} x_{J}
$$

Define, for each pair $I$ and $J$, the indexing sequence $((I, J))$ consisting of the increasing ordered union of the sequences $I$ and $J$. Moreover, for each $((I, J))$, there will be an associated sign

$$
\begin{equation*}
\operatorname{sign}(I, J):=\mid\{j-i: j-i<0 \text { for }(i, j) \in I \times J\} \mid \tag{A.53}
\end{equation*}
$$

that is, the number of times the indices of $J$ is greater than the ones from $I$. From the skew-commutativity property (Proposition 6.5.12),

$$
\begin{aligned}
\mathrm{d}(\omega \wedge \gamma) & =\sum_{I, J} \operatorname{sign}(I, J) \mathrm{d}\left(f_{I} \cdot g_{J}\right) \wedge \mathrm{d} x_{(I, J))} \\
& =\sum_{(I, J)} \operatorname{sign}(I, J)\left(\mathrm{d} f_{I} \cdot g_{J}+f_{I} \cdot \mathrm{~d} g_{J}\right) \wedge \mathrm{d} x_{((I, J))} \\
& =\sum_{I, J}\left(\mathrm{~d} f_{I} \cdot g_{J}+f_{I} \cdot \mathrm{~d} g_{J}\right) \wedge \mathrm{d} x_{I} \wedge \mathrm{~d} x_{J}
\end{aligned}
$$

From the distributive property, associativity and skew-commutativity we get

$$
\begin{aligned}
\mathrm{d}(\omega \wedge \gamma) & =\sum_{I, J}\left(g_{J} \mathrm{~d} f_{I} \wedge \mathrm{~d} x_{I} \wedge \mathrm{~d} x_{J}+f_{I} \mathrm{~d} g_{J} \wedge \mathrm{~d} x_{I} \wedge \mathrm{~d} x_{J}\right) \\
& =\sum_{I, J}\left(\mathrm{~d} f_{I} \wedge \mathrm{~d} x_{I}\right) \wedge\left(g_{J} \mathrm{~d} x_{J}\right)+(-1)^{k}\left(f_{I} \mathrm{~d} x_{I}\right) \wedge\left(\mathrm{d} g_{J} \wedge \mathrm{~d} x_{J}\right) \\
& =\sum_{I, J}\left(\mathrm{~d} f_{I} \wedge \mathrm{~d} x_{I}\right) \wedge\left(g_{J} \mathrm{~d} x_{J}\right)+(-1)^{k} \sum_{I, J}\left(f_{I} \mathrm{~d} x_{I}\right) \wedge\left(\mathrm{d} g_{J} \wedge \mathrm{~d} x_{J}\right) \\
& =\left(\sum_{I} \mathrm{~d} f_{I} \wedge \mathrm{~d} x_{I}\right) \wedge\left(\sum_{J} g_{J} \mathrm{~d} x_{J}\right)+(-1)^{k}\left(\sum_{I} f_{I} \mathrm{~d} x_{I}\right) \wedge\left(\sum_{J} \mathrm{~d} g_{J} \wedge \mathrm{~d} x_{J}\right) \\
& =\mathrm{d} \omega \wedge \lambda+(-1)^{k} \omega \wedge \mathrm{~d} \lambda
\end{aligned}
$$

(b) For the case of a zero form $f: E \rightarrow \mathbf{R}$, of class $C^{2}$, we have

$$
\mathrm{d}(\mathrm{~d} f)=\mathrm{d}\left(\sum_{j=1}^{n} \partial_{j} f \mathrm{~d} x_{j}\right)=\sum_{j=1}^{n} \mathrm{~d}\left(\partial_{j} f\right) \mathrm{d} x_{j}=\sum_{i, j=1}^{n} \partial_{i j} f \mathrm{~d} x_{i} \wedge \mathrm{~d} x_{j}
$$

Notice however that $\partial_{i j} f=\partial_{j i} f$ for any $1 \leqslant i, j \leqslant n$, thus each pair $(i, j)$ and $(j, i)$ of the sum cancel with each other - since $\mathrm{d} x_{i} \wedge \mathrm{~d} x_{j}=-\mathrm{d} x_{j} \wedge \mathrm{~d} x_{i}$ - which implies in $\mathrm{d}(\mathrm{d} f)=0$.
For the general case, if $\omega=\sum_{I} f_{I} \mathrm{~d} x_{I}$ is any $C^{2} k$-form, since $\mathrm{d}^{2} \omega=\sum_{I} \mathrm{~d}^{2} f_{I} \mathrm{~d} x_{I}$ and $\mathrm{d}^{2} f_{I}=0$, for all $I$, it follows immediately that $\mathrm{d}^{2} \omega=0$.

## Change of Variables - The Pullback Operation

Let $\omega:=\sum_{I} f_{I} \mathrm{~d} x_{I}$ be a $k$-form in an open set $E \subseteq \mathbf{R}^{n}$ and $\phi: V \rightarrow E$ be a $C^{1}$ map from another open set $V \subseteq \mathbf{R}^{m}$. Notice that there arises a natural pullback operation $\phi^{*}$ that
allows for a change of variables

$$
\begin{equation*}
\phi^{*}(\omega)=\sum_{I} f_{I} \phi \mathrm{~d} v_{I} \tag{A.54}
\end{equation*}
$$

where $v_{I}$ represents the coordinates coming from $V$.
Proposition A.10.11 (Pullback properties). Let $\omega$ and $\lambda$ be, respectively, a $k$-form and an $m$-form in the open set $E \subseteq \mathbf{R}^{n}$. Let $\phi: V \rightarrow E$ be a $C^{1}$ map. The following properties hold:
(a) If $k=m$, then $\phi^{*}(\omega+\lambda)=\phi^{*}(\omega)+\phi^{*}(\lambda)$.
(b) $\phi^{*}(\omega \wedge \lambda)=\phi^{*}(\omega) \wedge \phi^{*}(\lambda)$.
(c) If $\omega$ is $C^{1}$ and $\phi$ is of class $C^{2}$, then $\mathrm{d}\left(\phi^{*}(\omega)\right)=\phi^{*}(\mathrm{~d} \omega)$.

Proof. Let $\omega:=\sum f_{I} \mathrm{~d} x_{I}$ and $\lambda:=\sum g_{J} \mathrm{~d} x_{J}$.
(a) If $k=m$, then $\omega+\lambda$ is a $k$-form in $E$, and we have

$$
\phi^{*}(\omega+\lambda)=\phi^{*}\left(\sum f_{I} \mathrm{~d} x_{I}+\sum g_{J} \mathrm{~d} x_{J}\right)=\sum f_{I} \phi \mathrm{~d} v_{I}+\sum g_{J} \phi \mathrm{~d} v_{J}=\phi^{*}(\omega)+\phi^{*}(\lambda)
$$

(b) Notice that

$$
\begin{aligned}
\phi^{*}(\omega \wedge \lambda) & =\phi^{*}\left(\sum_{I, J} f_{I} \cdot g_{J} \mathrm{~d} x_{I} \wedge \mathrm{~d} x_{J}\right)=\sum_{I, J}\left(f_{I} \cdot g_{J}\right) \phi \mathrm{d} v_{I} \wedge \mathrm{~d} v_{J} \\
& =\sum_{I, J}\left(f_{I} \phi\right) \cdot\left(g_{J} \phi\right) \mathrm{d} v_{I} \wedge \mathrm{~d} v_{J}=\sum_{I, J}\left(f_{I} \phi \mathrm{~d} v_{I}\right) \wedge\left(g_{J} \phi \mathrm{~d} v_{J}\right) \\
& =\phi^{*}(\omega) \wedge \phi^{*}(\lambda)
\end{aligned}
$$

(c) We first prove the equality for the base case of a 0 -form. Let $h: E \rightarrow \mathbf{R}$ be a 0 -form, then

$$
\begin{aligned}
\phi^{*}(\mathrm{~d} h) & =\phi^{*}\left(\sum_{i=1}^{n} \partial_{i} h \mathrm{~d} x_{i}\right)=\sum_{i=1}^{n}\left(\partial_{i} h\right) \phi \mathrm{d} v_{i} \\
& =\sum_{i=1}^{n}\left(\partial_{i} h\right) \phi\left(\sum_{j=1}^{n} \partial_{j} v_{i} \mathrm{~d} x_{j}\right)=\sum_{j=1}^{n} \partial_{j}(h \phi) \mathrm{d} x_{j} \\
& =\mathrm{d}\left(\phi^{*}(h)\right)
\end{aligned}
$$

We now turn to the general case. Notice that $\phi^{*}(\omega)=\sum_{I} \phi^{*}\left(f_{I}\right) \mathrm{d} v_{I}=\sum_{I} \phi^{*}\left(f_{I}\right) \phi^{*}\left(\mathrm{~d} x_{I}\right)$, since $\phi^{*}\left(\mathrm{~d} x_{I}\right)=\mathrm{d} v_{I}$. Therefore, assuming $\phi$ is of class $C^{2}$, we have

$$
\begin{aligned}
\mathrm{d}\left(\phi^{*} \omega\right) & =\sum_{I} \mathrm{~d}\left(\phi^{*} f_{I}\right) \wedge \mathrm{d} v_{I}=\sum_{I} \phi^{*}\left(\mathrm{~d} f_{I}\right) \wedge \phi^{*}\left(\mathrm{~d} x_{I}\right) \\
& =\sum_{I} \phi^{*}\left(\mathrm{~d} f_{I} \wedge \mathrm{~d} x_{I}\right)=\phi^{*}\left(\sum_{I} \mathrm{~d} f_{I} \wedge \mathrm{~d} x_{I}\right) \\
& =\phi^{*}(\mathrm{~d} \omega)
\end{aligned}
$$

The need for $\phi$ to be $C^{2}$ comes from the fact that the first equality - by Theorem A.10.10 - is obtained by noting that

$$
\mathrm{d}\left(\phi^{*}(f) \mathrm{d} v_{I}\right)=\mathrm{d}\left(\phi^{*}(f)\right) \wedge \mathrm{d} v_{I}+\phi^{*}(f) \wedge \underbrace{\mathrm{d}\left(\mathrm{~d} v_{I}\right)}_{0}=\mathrm{d}\left(\phi^{*} f\right) \wedge \mathrm{d} v_{I}
$$

Lemma A.10.12 (Pullback composition). Let $\omega$ be a $k$-form in an open set $W \subseteq \mathbf{R}^{p}$. Consider two composable $C^{1}$ maps $\phi: E \rightarrow V$ and $\psi: V \rightarrow W$, where $U \subseteq \mathbf{R}^{n}$ and $V \subseteq \mathbf{R}^{m}$. Then the following equality holds

$$
\phi^{*}\left(\psi^{*}(\omega)\right)=(\psi \phi)^{*}(\omega) .
$$

Moreover, these are $k$-forms in $U$.
Proof. Define $\omega:=\sum_{I} f_{I} \mathrm{~d} w_{I}$. Thus

$$
\phi^{*}\left(\psi^{*}(\omega)\right)=\phi^{*}\left(\sum_{I} f_{I} \psi \mathrm{~d} v_{I}\right)=\sum_{I}\left(f_{I} \psi\right) \phi \mathrm{d} u_{I}=\sum_{I} f_{I}(\psi \phi) \mathrm{d} u_{I}=(\psi \phi)^{*}(\omega)
$$

Lemma A.10.13. Let $\omega$ be a $k$-form in an open set $E \subseteq \mathbf{R}^{n}$. We consider two $k$-surfaces $\Phi: D \rightarrow E$ and $\operatorname{id}_{D}: D \rightarrow \mathbf{R}^{k}$, with $\operatorname{id}_{D}(t):=t-$ where $D \subseteq \mathbf{R}^{k}$ is a compact set. Then

$$
\int_{\Phi} \omega=\int_{\mathrm{id}_{D}} \Phi^{*}(\omega)
$$

Proof. Define $\omega:=\sum_{I} f_{I} \mathrm{~d} x_{I}$, and $\Phi_{j}:=\pi_{j} \Phi$ as the $j$-th projection of the surface $\Phi$ that is

$$
\Phi^{*}(\omega)=\sum_{I} f_{I} \Phi \mathrm{~d} \Phi_{I}
$$

Defining $I:=\left(i_{p}\right)_{p=1}^{k}$, and since $\mathrm{d} \Phi_{i_{p}}=\sum_{q=1}^{k} \partial_{q} \Phi_{i_{p}} \mathrm{~d} t_{q}$ for all $1 \leqslant p \leqslant k$, we see that

$$
\begin{aligned}
\mathrm{d} \Phi_{i_{1}} \wedge \cdots \wedge \mathrm{~d} \Phi_{i_{k}} & =\left(\sum_{q=1}^{k} \partial_{q} \Phi_{i_{1}} \mathrm{~d} t_{q}\right) \wedge \cdots \wedge\left(\sum_{q=1}^{k} \partial_{q} \Phi_{i_{k}} \mathrm{~d} t_{q}\right) \\
& =\sum_{1 \leqslant i_{1}, \ldots, i_{k} \leqslant k}\left[\prod_{q=1}^{k} \partial_{q} \Phi_{i_{q}}\right] \mathrm{d} t_{i_{1}} \wedge \cdots \wedge \mathrm{~d} t_{i_{k}} \\
& =\sum_{1 \leqslant i_{1}, \ldots, i_{k} \leqslant k}\left[\prod_{q=1}^{k} \partial_{q} \Phi_{i_{q}}\right]\left(\operatorname{sign}\left(i_{1}, \ldots, i_{k}\right) \mathrm{d} t_{1} \wedge \cdots \wedge \mathrm{~d} t_{k}\right) \\
& =\operatorname{det}\left[\partial_{q} \Phi_{i_{p}}\right]_{p, q=1}^{n} \mathrm{~d} t_{1} \wedge \cdots \wedge \mathrm{~d} t_{k} \\
& =\operatorname{det}(\operatorname{Jac} \Phi) \mathrm{d} t_{1} \wedge \cdots \wedge \mathrm{~d} t_{k}
\end{aligned}
$$

Where the sign map is defined as in Eq. (A.53). Therefore, combining equations we find

$$
\Phi^{*}(\omega)=\sum_{I} f_{I} \Phi \operatorname{det}(\operatorname{Jac} \Phi) \mathrm{d} t_{1} \wedge \cdots \wedge \mathrm{~d} t_{k}
$$

thus indeed

$$
\begin{aligned}
\int_{\operatorname{id}_{D}} \Phi^{*}(\omega) & =\int_{\operatorname{id}_{D}} \sum_{I} f_{I} \Phi(t) \operatorname{det}(\operatorname{Jac} \Phi(t)) \mathrm{d} t_{1} \wedge \cdots \wedge \mathrm{~d} t_{k} \\
& =\int_{D} \sum_{I} f_{I} \Phi(t) \operatorname{det}(\operatorname{Jac} \Phi(t)) \mathrm{d} t \\
& =\int_{\Phi} \omega .
\end{aligned}
$$

Theorem A.10.14. Let $\psi: V \rightarrow E$ be a map of class $C^{1}$ on open sets $V \subseteq \mathbf{R}^{m}$, and $E \subseteq \mathbf{R}^{n}$. Consider a $k$-surface $\Phi: D \rightarrow V$, where $D \subseteq \mathbf{R}^{k}$ is compact, and a $k$-form $\omega$ in $E$. Then we have the following equality

$$
\int_{\psi \Phi} \omega=\int_{\Phi} \psi^{*}(\omega)
$$

Proof. The theorem is a direct consequence of the preceding lemmas - notice that

$$
\int_{\psi \Phi} \omega=\int_{\mathrm{id}_{D}}(\psi \Phi)^{*} \omega=\int_{\mathrm{id}_{D}} \Phi^{*}\left(\psi^{*}(\omega)\right)=\int_{\Phi} \psi^{*}(\omega)
$$

## Simplexes and Chains

Notation A.10.15. For the purposes of this section, we are going to define the collection

$$
\Delta_{\mathrm{c}}^{k}:=\left\{\left(t_{1}, \ldots, t_{k}\right) \in \mathbf{R}^{k}: \sum_{j=1}^{k} t_{j} \leqslant 1, \text { and } t_{j} \geqslant 0 \text { for all } 1 \leqslant j \leqslant k\right\}
$$

which represents the $k$-simplex obtained by the corner of a $k$-dimensional unit cube.
Definition A.10.16 (Affine map). Given vector spaces $V$ and $L$, we say that $f: V \rightarrow L$ is a affine map if $f(x)=\ell+\phi(x)$ for all $x \in V$, where $\phi: V \rightarrow L$ is a $k$-linear map and $\ell=f(0) \in L$.

Definition A.10.17 (Oriented affine $k$-simplex). Let $p_{0}, \ldots, p_{k} \in \mathbf{R}^{n}$ be any points. We define the oriented affine $k$-simplex induced by the points $p_{0}, \ldots, p_{k}$ in $\mathbf{R}^{n}$ as an affine $\operatorname{map} \sigma: \Delta_{\mathrm{c}}^{k} \rightarrow \mathbf{R}^{n}$ given by

$$
\sigma(\alpha):=p_{0}+\sum_{j=1}^{k} \alpha_{j}\left(p_{j}-p_{0}\right), \text { for all } \alpha \in \Delta_{\mathrm{c}}^{k}
$$

We can also denote $\sigma$ by $\left[p_{0}, \ldots, p_{k}\right]$ as in Definition 6.6.12. If we define the $\mathbf{R}$-linear $\operatorname{map} A: \mathbf{R}^{k} \rightarrow \mathbf{R}^{n}$ to be given by $A\left(e_{j}\right):=p_{j}-p_{0}$, then we can write $\sigma(\alpha)=p_{0}+A(\alpha)$.

If $\tau \in S_{k+1}$ is a permutation on $k+1$ elements, and $\mu:=\left[p_{\tau(0)}, p_{\tau(1)}, \ldots, p_{\tau(k)}\right]$ is an oriented affine $k$-simplex, also induced by the points $p_{0}, \ldots, p_{k}$, we say that

$$
\tau=\operatorname{sign}(\tau) \sigma .
$$

Moreover, $\mu$ is said to have the same orientation of $\sigma$ if $\operatorname{sign}(\tau)>0$, otherwise, if $\operatorname{sign}(\tau)<0$, then $\mu$ is said to have the opposite orientation of $\sigma$.

In the special case where $k=n$ and the collection $\left\{p_{j}-p_{0}\right\}_{j=1}^{k}$ is linearly independent, we say that:

- $\sigma$ is positively oriented if $\operatorname{det} A>0$.
- $\sigma$ is negatively oriented if $\operatorname{det} A<0$.

In particular, the simplex $\mathrm{id}_{\mathbf{R}^{k}}:=\left[0, e_{1}, \ldots, e_{k}\right]$ in $\mathbf{R}^{k}$ is positively oriented.
Example A.10.18. A special example of an oriented affine $k$-simplex occurs for the case $k=0$, where we get a simplex induced by a single point $p_{0} \in \mathbf{R}^{k}$ - in such case, two 0 -simplexes are conceivable, $\sigma=p_{0}$ or $\sigma=-p_{0}$.

Definition A.10.19. Let $\sigma=\varepsilon p_{0}$ be an oriented affine 0 -simplex - where $\varepsilon \in\{-1,1\}$ and $p_{0} \in \mathbf{R}^{k}$. If $f: E \rightarrow \mathbf{R}$ is a 0 -form in the open set $E \subseteq \mathbf{R}^{n}$, we define its integral over the 0 -simplex $\sigma$ as

$$
\int_{\sigma} f:=\varepsilon f\left(p_{0}\right)
$$

Proposition A.10.20. Let $\sigma: \Delta_{\mathrm{c}}^{k} \rightarrow E$ be an oriented affine $k$-simplex in an open set $E \subseteq \mathbf{R}^{n}$, and $\omega$ be any $k$-form in $E$. If $\mu=\varepsilon \sigma$ where $\varepsilon \in\{-1,1\}^{2}$, then

$$
\int_{\mu} \omega=\varepsilon \int_{\sigma} \omega
$$

Proof. For the base case $k=0$, we have Definition A.10.19. Now, let $k \geqslant 1$ and define, for each $1 \leqslant j_{0} \leqslant k$, the transposition $\tau_{j_{0}} \in S_{k+1}$ given by $\tau_{j_{0}}(0)=j_{0}$ and $\tau_{j_{0}}\left(j_{0}\right)=0$ while $\tau_{j_{0}}(i)=i$ for all $i \neq 0, j_{0}$. Suppose $\sigma:=\left[p_{0}, \ldots, p_{k}\right]$, then since $\operatorname{sign}\left(\tau_{j_{0}}\right)=-1$, the oriented affine $k$-simplex $\mu:=\left[p_{\tau_{0}}(0), \ldots, p_{\tau_{j_{0}}(k)}\right]$ is such that $\mu=\operatorname{sign}\left(\tau_{j_{0}}\right) \sigma=-\sigma-$ moreover,

$$
\mu(u)=p_{j_{0}}+\sum_{j=1}^{k} u_{i}\left(p_{\tau_{j 0}(i)}-p_{j_{0}}\right):=p_{j_{0}}+B(u), \text { for all } u \in \Delta_{\mathrm{c}}^{k}
$$

where $B: \mathbf{R}^{k} \rightarrow \mathbf{R}^{n}$ is a linear map with matrix representation $B=\left[p_{\tau_{j_{0}}}(j)-p_{j_{0}}\right]_{j=1}^{k}$ (notice we are disregarding the column with $j=0$ ), where each $p_{\tau_{0}(j)}-p_{j_{0}}=\left[\partial_{j} \mu_{i}\right]_{i=1}^{n}$

[^41]is the $j$-th column of $B$. Therefore, if $\omega:=\sum_{I} f_{I} \mathrm{~d} x_{I}$ and $I:=\left(i_{p}\right)_{p=1}^{k}$ for each $I$, we obtain
$$
\int_{\mu} \omega=\int_{\Delta_{\mathrm{c}}^{k}} \sum_{I} f_{I} \mu(t) \operatorname{det}(B) \mathrm{d} t
$$

If we let $A: \mathbf{R}^{k} \rightarrow \mathbf{R}^{n}$ be the linear map defined by $A(u):=\sum_{i=1}^{k} u_{i}\left(p_{i}-p_{0}\right)$ for all $u \in \mathbf{R}^{k}$, we can write $\sigma(u)=p_{0}+A(u)$. Notice however that, for each $1 \leqslant i \leqslant k$ with $i \neq j_{0}$, we have the relation $B\left(e_{i}\right)=A\left(e_{i}\right)-A\left(e_{j_{0}}\right)$ (which does not affect the value of the determinant) - while $B\left(e_{j_{0}}\right)=-A\left(e_{j_{0}}\right)$ (which multiplies the determinant by -1 ). Therefore $\operatorname{det} B=-\operatorname{det} A$. Hence,

$$
\begin{equation*}
\int_{\mu} \omega=\int_{\Delta_{\mathrm{c}}^{k}} \sum_{I} f_{I} \mu(t)(-\operatorname{det} A) \mathrm{d} t=\int_{\Delta_{\mathrm{c}}^{k}} \sum_{I} f \sigma(t) \operatorname{det}(A) \mathrm{d} t=\int_{\sigma} \omega . \tag{A.55}
\end{equation*}
$$

For the case where we transpose a pair of indices $i$ and $j$, with $0<i<j \leqslant k$, then $\mu(u)=p_{0}+C(u)$ - where $C$ has all columns matching to those of $A$, except the interchange between the $i$-th and $j$-th columns, which implies in $\operatorname{det} C=-\operatorname{det} A$. Therefore we get again the same as in Eq. (A.55), which finishes the proof.

Definition A.10.21 (Affine $k$-chain). Let $E \subseteq \mathbf{R}^{n}$ be an open set. A finite collection of oriented affine $k$-simplexes in $E$ is said to be an affine $k$-chain in $E$. It is to be noted that the finite collection may contain some multiplicities - that is, some of the $k$-simplexes can coincide.

If $\Gamma:=\left(\sigma_{j}\right)_{j=1}^{m}$ is an affine $k$-chain in $E$, we define

$$
\int_{\Gamma} \omega:=\sum_{j=1}^{m} \int_{\sigma_{j}} \omega
$$

We'll usually denote $\Gamma$ by the formal sum $\sum_{j=1}^{m} \sigma_{j}$ mapping $\omega \stackrel{\Gamma}{\mapsto} \sum_{j} \int_{\sigma_{j}} \omega$.
Definition A.10.22 (Boundary). Let $k \geqslant 1$. Given an oriented affine $k$-simplex $\sigma:=$ [ $p_{0}, \ldots, p_{k}$ ], we define the boundary of $\sigma$ to be an affine $(k-1)$-chain defined by

$$
\partial \sigma:=\sum_{j=0}^{k}(-1)^{j}\left[p_{0}, \ldots, p_{j-1}, p_{j+1}, \ldots, p_{k}\right] .
$$

Definition A.10.23 (Differentiable simplexes and chains). Let $E \subseteq \mathbf{R}^{n}$ and $V \subseteq \mathbf{R}^{m}$ be open sets. If $\phi: E \rightarrow V$ is a map of class $C^{2}$, and $\sigma$ is an oriented affine $k$-simplex in $E$, then the induced map

$$
\Phi:=\phi \sigma: \Delta_{\mathrm{c}}^{k} \longrightarrow V
$$

is a $k$-surface in $V$. The surface $\Phi$ is said to be an oriented $k$-simplex of class $C^{2}$ in $V$. Moreover, $\Phi$ has a boundary defined by

$$
\partial \Phi:=\phi(\partial \sigma),
$$

therefore, $\partial \Phi$ is a $(k-1)$-chain of class $C^{2}$ in $V$.
A finite collection $\Psi:=\left(\Phi_{j}\right)_{j=1}^{r}$ of oriented $k$-simplexes of class $C^{2}$ in $V$ is said to be a $k$-chain of class $C^{2}$ in $V$. Given a $k$-form $\omega$ in $V$, we define

$$
\int_{\Psi} \omega:=\sum_{j=1}^{r} \int_{\Phi_{j}} \omega
$$

As expected, the boundary of $\Psi$ is defined to be the $(k-1)$-chain of class $C^{2}$ in $V$ given by $\partial \Psi:=\sum_{j=1}^{r} \partial \Phi_{j}$.

In the context of the last definition, given an affine $k$-chain $\Gamma:=\sum_{j=1}^{r} \sigma_{j}$ in $E$, the map $\phi$ induces a corresponding $k$-chain of class $C^{2}$ in $V$, namely, $\sum_{j=1}^{r} \phi \sigma_{j}$.

## Oriented Boundaries on Sets

Definition A.10.24 (Positively oriented boundaries on a set). Let $\phi: \Delta_{c}^{n} \mapsto \mathbf{R}^{n}$ be an injective map of class $C^{2}$, whose Jacobian determinant is positive in the interior of $\Delta_{\mathrm{c}}^{k}$. Let $E:=\phi\left(\Delta_{\mathrm{c}}^{n}\right)$ — then, by the inverse map theorem (see Theorem A.5.8), $E$ is the closure of an open set of $\mathbf{R}^{n}$. We define the positively oriented boundary of the set $E$ to be the $(n-1)$-chain

$$
\partial E:=\partial \phi=\phi\left(\partial \mathrm{id}_{\mathbf{R}^{n}}\right),
$$

where $\operatorname{id}_{\mathbf{R}^{n}}:=\left[0, e_{1}, \ldots, e_{n}\right]$ - that is to say that, for any $(n-1)$-form $\omega$ in $E$, we have

$$
\int_{\partial E} \omega:=\int_{\partial \phi} \omega
$$

If $\left\{E_{j}\right\}_{j=1}^{r}$ is a collection of subsets of $\mathbf{R}^{n}$ with disjoint interior, let $\left\{\phi_{j}: \Delta_{\mathrm{c}}^{n} \mapsto \mathbf{R}^{n}\right\}_{j=1}^{r}$ be an associated collection of injective $C^{2}$-maps with positive Jacobian determinant in the interior of $\Delta_{\mathrm{c}}^{n}$, and such that $\phi_{j}\left(\Delta_{\mathrm{c}}^{n}\right)=E_{j}$. We define the positively oriented boundary of the set $\Omega:=E_{1} \cup \cdots \cup E_{r}$ as the ( $n-1$ )-chain

$$
\partial \Omega:=\partial \phi_{1}+\cdots+\partial \phi_{r}, \text { that is } \int_{\partial \Omega} \omega:=\sum_{j=1}^{r} \int_{\partial \phi_{j}} \omega,
$$

where $\omega$ is any $(n-1)$-form in $\Omega$.
Proposition A.10.25. Let $\phi, \psi: \Delta_{\mathrm{c}}^{n} \rightrightarrows \mathbf{R}^{n}$ be injective $C^{2}$-maps with positive Jacobian determinant in the interior of $\Delta_{\mathrm{c}}^{n}$. If $\phi\left(\Delta_{\mathrm{c}}^{n}\right)=\psi\left(\Delta_{\mathrm{c}}^{n}\right)$, then $\partial \phi=\partial \psi$ - that is, for every ( $n-1$ )-form $\omega$ in the image of the maps, we have

$$
\int_{\partial \phi} \omega=\int_{\partial \psi} \omega
$$

Proof.
Prove

Example A.10.26. Let $S:[0, \pi] \times[0,2 \pi] \rightarrow \mathbf{R}^{3}$ be a 2 -surface in $\mathbf{R}^{3}$, defined by

$$
S(u, v):=(\sin (u) \cos (v), \sin (u) \sin (v), \cos (u)) .
$$

Notice that the positively oriented boundary of $S$ corresponds to the 4 distinct curves induced by $S$ under the positively oriented boundary of the rectangle $[0, \pi] \times[0,2 \pi]$ - that is, $\partial S=\sum_{j=1}^{4} \gamma_{j}$ where we have

$$
\begin{aligned}
& \gamma_{1}:[0, \pi] \longrightarrow \mathbf{R}^{3} \text { mapping } u \longmapsto S(u, 0)=(\sin (u), 0, \cos (u)) ; \\
& \gamma_{2}:[0,2 \pi] \longrightarrow \mathbf{R}^{3} \text { mapping } v \longmapsto S(\pi, v)=(0,0,-1) ; \\
& \gamma_{3}:[0, \pi] \longrightarrow \mathbf{R}^{3} \text { mapping } u \longmapsto S(\pi-u, 2 \pi)=(\sin (u), 0,-\cos (u)) ; \\
& \gamma_{4}:[0,2 \pi] \longrightarrow \mathbf{R}^{3} \text { mapping } v \longmapsto S(0,2 \pi-v)=(0,0,1) .
\end{aligned}
$$

Now, given any 1-form $\omega:=f_{1} \mathrm{~d} x_{1}+f_{2} \mathrm{~d} x_{2}+f_{3} \mathrm{~d} x_{3}$ in $\mathbf{R}^{3}$, we have

$$
\begin{aligned}
\int_{\partial S} \omega & =\sum_{j=1}^{4} \int_{\gamma_{j}} \omega \\
& =\int_{0}^{\pi} \sum_{i=1}^{3} f_{i} \gamma_{1}(x) \underbrace{\partial \gamma_{1}^{(i)}(x)}_{0} \mathrm{~d} x+\int_{\gamma_{2}} \omega+\int_{\gamma_{3}} \omega+\int_{0}^{2 \pi} \sum_{i=1}^{3} f_{i} \gamma_{4}(x) \underbrace{\partial \gamma_{4}^{(i)}(x)}_{0} \mathrm{~d} x \\
& =\int_{\gamma_{2}} \omega+\int_{\gamma_{3}} \omega .
\end{aligned}
$$

Notice however that if we make the change of variables from $x$ to $\pi-x$ in the integral $\int_{\gamma_{3}} \omega$, we obtain

$$
\begin{aligned}
\int_{\gamma_{3}} \omega & =\int_{0}^{\pi}-f_{1} \gamma_{3}(x) \cos (x)+f_{3} \gamma_{3}(x) \sin (x) \mathrm{d} x \\
& =-\int_{\pi}^{0}-f_{1} \gamma_{3}(\pi-x) \cos (x)+f_{3} \gamma_{3}(\pi-x) \sin (x) \mathrm{d} x \\
& =\int_{0}^{\pi}-f_{1} \gamma_{2}(x) \cos (x)+f_{3} \gamma_{2}(x) \sin (x) \mathrm{d} x \\
& =-\int_{\gamma_{2}} \omega
\end{aligned}
$$

Where the third equality may be obtained by noting that $\gamma_{3}(\pi-x)=\gamma_{2}(x)$. We conclude that

$$
\int_{\partial S} \omega=0
$$

and, more generally, $\partial S=0$.

## Stoke's Theorem

Theorem A.10.27 (Stoke's). Let $\Psi$ be a $k$-chain of class $C^{2}$ in an open set $V \subseteq \mathbf{R}^{n}$, and let $\omega$ be a $(k-1)$-form of class $C^{1}$ in $V$. Then the following equality holds

$$
\int_{\Psi} \mathrm{d} \omega=\int_{\partial \Psi} \omega .
$$

Proof. Since $\Psi=\sum_{j=1}^{r} \Phi_{j}$ for some finite collection $\left(\Phi_{j}\right)_{j=1}^{r}$ of oriented $k$-simplexes of class $C^{2}$, it suffices to prove that the theorem holds for any oriented $k$-simplex of class $C^{2}$, since $\int_{\Psi} \gamma=\sum_{j=1}^{r} \int_{\Phi_{j}} \gamma$ for any $(k-1)$-form $\gamma$ in $V$.

Let $\Phi$ be any oriented $k$-simplex of class $C^{2}$ in $V$, and consider the positively oriented affine $k$-simplex $\operatorname{id}_{\mathbf{R}^{k}}:=\left[0, e_{1}, \ldots, e_{k}\right]: \Delta_{\mathrm{c}}^{k} \rightarrow \mathbf{R}^{k}$. From Definition A.10.23, there must exist an open set $E \subseteq \mathbf{R}^{k}$ with $\Delta_{\mathrm{c}}^{k} \subseteq E$, and a $C^{2}$-map $\phi: E \rightarrow V$ such that $\Phi=\phi \operatorname{id}_{\mathbf{R}^{k}}$. Therefore, given any $(k-1)$-form $\omega$ in $V$, we have

$$
\int_{\Phi} \mathrm{d} \omega=\int_{\phi \mathrm{id}_{\mathbf{R}^{k}}} \mathrm{~d} \omega=\int_{\mathrm{id}_{\mathbf{R}^{k}}} \phi^{*}(\mathrm{~d} \omega)=\int_{\mathrm{id}_{\mathrm{R}^{k}}} \mathrm{~d}\left(\phi^{*}(\omega)\right) .
$$

Moreover, we can also consider the positively oriented boundary of $\Phi$, which yields

$$
\int_{\partial \Phi} \omega=\int_{\partial\left(\phi \mathrm{id}_{\mathbf{R}^{k}}\right)} \omega=\int_{\phi\left(\partial \mathrm{id}_{\mathbf{R}^{k}}\right)} \omega=\int_{\partial \mathrm{id}_{\mathbf{R}^{k}}} \phi^{*}(\omega) .
$$

Hence, our goal will be to prove that the theorem is true for any ( $k-1$ )-form of the type $\phi^{*}(\omega)$, that is, any $(k-1)$-form $\gamma$ in $E$ it should be true that

$$
\begin{equation*}
\int_{\mathrm{id}_{\mathrm{R}^{k}}} \mathrm{~d} \gamma=\int_{\partial \mathrm{id}_{\mathrm{R}^{k}}} \gamma \tag{A.56}
\end{equation*}
$$

Let's first deal with the base case, $k=1$. Given any 0 -form $\gamma: E \rightarrow \mathbf{R}$ of class $C^{1}$ in $E \subseteq \mathbf{R}$, we have - by the fundamental theorem of calculus,

$$
\int_{\mathrm{id}_{\mathrm{R}}} \mathrm{~d} \gamma=\int_{0}^{1}(\mathrm{~d} \gamma(x)) \mathrm{d} x=\gamma(1)-\gamma(0)=\int_{\partial \mathrm{id}_{\mathrm{R}}} \gamma
$$

We now assume that $k>1$. Since any $(k-1)$-form $\gamma$ in $E$ can be written as a sum

$$
\gamma=\sum_{r=1}^{k} f_{r} \mathrm{~d} x_{1} \wedge \cdots \wedge \mathrm{~d} x_{r-1} \wedge \mathrm{~d} x_{r+1} \wedge \cdots \wedge \mathrm{~d} x_{k}:=\sum_{r=1}^{k} \gamma_{r}
$$

we may as well simply prove Eq. (A.56) for every $\gamma_{r}$, with $1 \leqslant r \leqslant k$ — this is now what we'll do, for each index $r$. The boundary of $\operatorname{id}_{\mathbf{R}^{k}}$ is given by

$$
\begin{equation*}
\partial \operatorname{id}_{\mathbf{R}^{k}}=\left[e_{1}, \ldots, e_{k}\right]+\sum_{j=1}^{k}(-1)^{j} \sigma_{j}, \tag{A.57}
\end{equation*}
$$

where we define each $\sigma_{j}$ to be the oriented ( $k-1$ )-simplex $\sigma_{j}:=\left[0, e_{1}, \ldots, e_{j-1}, e_{j+1}, \ldots, e_{k}\right]$ - that is, removing the $j$-th component of $\mathrm{id}_{\mathbf{R}^{k}}$, for each $1 \leqslant j \leqslant k$. Moreover, to facilitate our analysis, we'll define another oriented ( $k-1$ )-simplex, $\sigma_{0}$, defined by transposing the $r$-th component of $\mathrm{id}_{\mathbf{R}^{k}}$ an amount of $r-1$ times to the left, so that

$$
\sigma_{0}:=\left[e_{r}, e_{1}, \ldots, e_{r-1}, e_{r+1}, \ldots, e_{k}\right]=(-1)^{r-1}\left[e_{1}, \ldots, e_{k}\right]
$$

Therefore, we can rewrite Eq. (A.57) as

$$
\partial \operatorname{id}_{\mathbf{R}^{k}}=(-1)^{r-1} \sigma_{0}+\sum_{j=1}^{k}(-1)^{j} \sigma_{j} .
$$

We can analyse each of the $(k-1)$-simplexes $\sigma_{j}: \Delta_{\mathrm{c}}^{k-1} \rightarrow \mathbf{R}^{k}$, for $0 \leqslant j \leqslant k$ :

- For the special case $j=0$, given any $t \in \Delta_{\mathrm{c}}^{k-1}$, if $x:=\sigma_{0}(t) \in \mathbf{R}^{k}$, then

$$
x_{i}= \begin{cases}t_{i}, & 1 \leqslant i<r  \tag{A.58}\\ 1-\sum_{i=1}^{k-1} t_{i}, & i=r \\ t_{i-1}, & r<i \leqslant k\end{cases}
$$

Therefore, the matrix $\left[\partial_{q} \sigma_{0}^{(p)}(t)\right]_{i, j}$, where $1 \leqslant p, q \leqslant k-1$ and $p \neq r$, corresponds to the identity matrix on $\mathbf{R}^{k-1}$ restricted to $\Delta_{\mathrm{c}}^{k}$, therefore $\operatorname{det}\left[\partial_{q} \sigma_{0}^{(p)}\right]=1$ and thus

$$
\begin{equation*}
\int_{\sigma_{0}} \gamma_{r}=\int_{\Delta_{\mathrm{c}}^{k-1}} f_{r} \sigma_{r}(t) \mathrm{d} t \tag{A.59}
\end{equation*}
$$

- For any $j \neq 0$, given any $t \in \Delta_{\mathrm{c}}^{k}$, if $x:=\sigma_{j}(t) \in \mathbf{R}^{k}$, then

$$
x_{i}= \begin{cases}t_{i}, & 1 \leqslant i<j  \tag{A.60}\\ 0, & i=j \\ t_{i-1}, & j<i \leqslant k\end{cases}
$$

Moreover, since

$$
\int_{\sigma_{j}} \gamma_{r}=\int_{\Delta_{\mathrm{c}}^{k}} f_{r} \sigma_{j}(t) \operatorname{det}\left[\partial_{q} \sigma_{j}^{(p)}(t)\right]_{\substack{p, q=1 \\ p \neq r}}^{k-1} \mathrm{~d} t
$$

there are two possible cases, for any $t \in \Delta_{c}^{k}$ : if $j \neq r$, then the $j$-th row of the Jacobian matrix is entirely composed of zeros, hence $\operatorname{det}\left[\partial_{q} \sigma_{j}^{(p)}(t)\right]_{p, q}=0$, otherwise, if $j=r$, the matrix corresponds exactly to the identity matrix of $\mathbf{R}^{k-1}$ restricted to $\Delta_{\mathrm{c}}^{k-1}$ — therefore $\operatorname{det}\left[\partial_{q} \sigma_{j}^{(p)}(t)\right]_{p, q}=1$ for any $t \in \Delta_{\mathrm{c}}^{k-1}$, which implies in

$$
\int_{\sigma_{j}} \gamma_{r}= \begin{cases}0, & j \neq r  \tag{A.61}\\ \int_{\Delta_{\mathrm{c}}^{k-1}} f_{r} \sigma_{r}(t) \mathrm{d} t, & j=r\end{cases}
$$

With Eq. (A.59) and Eq. (A.61) at hand, we obtain

$$
\begin{aligned}
\int_{\partial \mathrm{id}_{\mathbf{R}^{k}}} \gamma_{r} & =(-1)^{r-1} \int_{\sigma_{0}} \gamma_{r}+\sum_{j=1}^{r}(-1)^{j} \int_{\sigma_{j}} \gamma_{r} \\
& =(-1)^{r-1} \int_{\sigma_{0}} \gamma_{r}+(-1)^{r} \int_{\sigma_{r}} \gamma_{r} \\
& =(-1)^{r-1} \int_{\Delta_{\mathrm{c}}^{k-1}} f_{r} \sigma_{0}(t)-f_{r} \sigma_{r}(t) \mathrm{d} t
\end{aligned}
$$

Notice, on the other hand, that

$$
\begin{aligned}
\mathrm{d} \gamma_{r} & =\mathrm{d} f_{r} \wedge \mathrm{~d} x_{1} \wedge \cdots \wedge \mathrm{~d} x_{r-1} \wedge \mathrm{~d} x_{r+1} \wedge \cdots \wedge \mathrm{~d} x_{k} \\
& =\left(\sum_{j=1}^{k} \partial_{j} f_{r} \mathrm{~d} x_{j}\right) \wedge \mathrm{d} x_{1} \wedge \cdots \wedge \mathrm{~d} x_{r-1} \wedge \mathrm{~d} x_{r+1} \wedge \cdots \wedge \mathrm{~d} x_{k} \\
& =\partial_{r} f_{r} \mathrm{~d} x_{r} \wedge \mathrm{~d} x_{1} \wedge \cdots \wedge \mathrm{~d} x_{r-1} \wedge \mathrm{~d} x_{r+1} \wedge \cdots \wedge \mathrm{~d} x_{k} \\
& =(-1)^{r-1} \partial_{r} f_{r} \mathrm{~d} x_{1} \wedge \cdots \wedge \mathrm{~d} x_{k} .
\end{aligned}
$$

Therefore, by Fubini's theorem (see Theorem A.8.4),

$$
\begin{aligned}
\int_{\mathrm{id}_{\mathbf{R}^{k}}} \mathrm{~d} \gamma & =(-1)^{r-1} \int_{\Delta_{\mathrm{c}}^{k}} \partial_{r} f_{r}(x) \mathrm{d} x \\
& =(-1)^{r-1} \int_{\Delta_{\mathrm{c}}^{k-1}}\left[\int_{0}^{1-x_{1}+\cdots+x_{r-1}+x_{r+1}+x_{k-1}} \partial_{r} f_{r}(x) \mathrm{d} x_{r}\right] \mathrm{d}\left(x_{1}, \ldots, x_{r-1}, x_{r+1}, \ldots, x_{k}\right) \\
& =(-1)^{r-1} \int_{\Delta_{\mathrm{c}}^{k-1}} f_{r} \sigma_{0}(t)-f_{r} \sigma_{r}(t) \mathrm{d} t
\end{aligned}
$$

Where the last equality was obtained by means of Eq. (A.58) and Eq. (A.60) since the variables are independent of each other. That is, we just proved that

$$
\int_{\mathrm{id}_{\mathbf{R}^{k}}} \mathrm{~d} \gamma_{r}=\int_{\partial \mathrm{id}_{\mathbf{R}^{k}}} \gamma_{r},
$$

which implies in Eq. (A.56) - thus the proof is complete!

## Appendix B

## Set Theory

## B. 1 ZFC

Theorem B.1.1 (Cantor). Let $A$ be a set and $f: A \rightarrow 2^{A}$ be a map, then $f$ is not surjective and hence $|A|<\left|2^{A}\right|$.

Proof. Let $B:=\{x \in A: x \notin f(x)\}$ be a set of $2^{A}$ and suppose, for the sake of contradiction, that $f$ is surjective. That is, there must exist some $y \in A$ for which $f(y)=B$ - notice however that this can't be the case since by the law of excluded middle if $y \in B$ then $y \notin f(y)$ thus $f(y) \neq B$, while if $y \notin B$ then $y \in f(y)$ and thus $f(y) \neq B$. Therefore $f$ cannot be surjective and thus $|A|<\left|2^{A}\right|$ since there exists an injective mapping $x \mapsto\{x\}$.

Theorem B.1.2. There is no set containing all sets as members.
Proof. For the sake of contradiction, let $A$ be a set containing every set as a member. In particular $2^{A} \subseteq A$ so that $\left|2^{A}\right| \leqslant A$ - this can't be the case by Theorem B.1.1, thus $A$ cannot be a set.

Definition B.1.3 (Cardinal numbers). A cardinal number is an isomorphism class of sets, and the cardinality of a given set $S$ is its isomorphism class. The following are properties pertaining to cardinal numbers:
(a) Every set has a unique cardinal number as its cardinality.
(b) Every cardinal number is the cardinality of some set.
(c) Two sets have the same cardinality if and only if they are isomorphic as sets.

Definition B.1.4 (Well ordering). A well-ordering on a set $S$ is a total ordering such that every non-empty subset of $S$ has a least element. A set equipped with a well-order is called a poset.

Definition B.1.5 (Ordinal). An ordinal number is an isomorphism class of well-ordered sets, and the ordinal rank of a poset $S$ is its isomorphism class. The following are properties satisfied by the ordinal numbers in its regard to ordinal ranks:
(a) Every poset has a unique ordinal numbers as its ordinal rank.
(b) Every ordinal number is the ordinal rank of some poset.
(c) Two sets have the same ordinal rank if and only if they are isomorphic as posets.

Moreover, ordinals also have the following properties:
(d) Every ordinal $\alpha$ has an immediate successor $\alpha+1$ - this process entails the addition of an element to the end of a chain of a well-ordering of type $\alpha$.
(e) There is a natural well-ordering on the collection of all ordinals. Given two ordinals $\alpha$ and $\beta$, we say that $\alpha \leqslant \beta$ if and only if there exists an initial segment of $\beta$ for which $\alpha$ is isomorphic to.
(f) The induced well-ordering on the set $\{\beta: \beta<\alpha\}$ is the isomorphism class represented by $\alpha$ - therefore, one can define an ordinal as a set containing all the smaller ordinals as members.
(g) (Equivalent to the Axiom of Choice) Every set is well-orderable, hence bijective to some ordinal.

We denote by $\omega$ the natural numbers, so that every ordinal bijective to $\omega$ is said to be countable.

The cardinal numbers are well-ordered, and can be indexed by means of ordinals. We denote the $\alpha$-th cardinal number by $\boldsymbol{\aleph}_{\alpha}$ — hence $\boldsymbol{\aleph}_{0}=\omega$, while $\boldsymbol{\aleph}_{1}=\omega_{1}$ (the first uncountable ordinal), for instance.

Definition B.1.6 (Successor \& limit ordinal). Given an ordinal $\alpha$, we can classify it as a successor ordinal, if there exists an ordinal $\beta$ such that $\alpha=\beta+1$, or as a limit ordinal in particular, every cardinal is a limit ordinal.

Definition B.1.7 (Successor \& limit cardinal). A cardinal is said to be a successor cardinal if its corresponding indexing ordinal is a successor ordinal, otherwise it is a limit cardinal.

## Bibliography

[ADH21] Araminta Amabel, Arun Debray, and Peter J. Haine. Differential Cohomology: Categories, Characteristic Classes, and Connections. 2021.
[Alu09] Paolo Aluffi. Algebra: Chapter 0. 2009.
[BBT20] Tai-Danae Bradley, Tyler Bryson, and John Terilla. Topology: A Categorical Approach. 2020.
[Bis06] Christopher Bishop. Pattern Recognition and Machine Learning. 2006. url: https://www.microsoft.com/en-us/research/publication/pattern-recognition-machine-learning/.
[Die16] Reinhard Diestel. Graph Theory. 2016.
[Eng89] Ryszard Engelking. General Topology. 1989.
[Jos06] Jürgen Jost. Postmodern Analysis. 2006.
[Kim20] Dongryul Kim. A rough guide to linear algebra. 2020. url: https://stanford. edu/~dkim04/assets/pdf/writings/LinAlg.pdf.
[KS06] Masaki Kashiwara and Pierre Schapira. Categories and Sheaves. 2006.
[Lan93] Serge Lang. Algebra. 1993.
[Lee11] John Lee. Introduction to Topological Manifolds. 2011.
[MK89] Yu I Manin and Alexei I Kostrikin. Linear Algebra and Geometry. 1989.
[Mun00] James Munkres. Topology. 2000.
[Rie16] Emily Riehl. Category Theory in Context. 2016.
[Rud76] Walter Rudin. Principles of Mathematical Analysis. 1976.
[SB14] Shai Shalev-Shwartz and Shai Ben-David. Understanding Machine Learning: From Theory to Algorithms. 2014. url: https://www. cs.huji. ac.il/ w~shais/UnderstandingMachineLearning.
[Zor15] Vladmir A. Zorich. Mathematical Analysis I. 2015.
[Zor16] Vladmir A. Zorich. Mathematical Analysis II. 2016.


[^0]:    ${ }^{1}$ We shall adopt the former notation, which should not cause any confusion.
    ${ }^{2}$ It should be stressed that this collection may well be empty.

[^1]:    ${ }^{3}$ When need be, we may use the symbol $g \circ f$ to denote the composition $g f$.

[^2]:    ${ }^{4}$ When convenient, we may discard the use of parenthesis, but in occasions where the use of parenthesis brings more clarity to the situation, we shall use it.

[^3]:    ${ }^{5}$ We say that $F_{0}$ is unique up to unique isomorphism when, given another functor $G: C \rightarrow D_{0}$ and natural isomorphism $\beta: F \xlongequal{\Longrightarrow} \iota G$, there exists a unique natural isomorphism $\eta: G \xlongequal{\Longrightarrow} F_{0}$ for which $\alpha=\iota \eta \beta$.
    ${ }^{6}$ A diagram whose nodes are categories an arrows are morphisms is said to be quasi-commutative if it commutes up to natural isomorphism of functors.

[^4]:    ${ }^{7}$ Recall Definition 1.4.21.

[^5]:    ${ }^{1}$ This is clearly an abuse of language, the equalizer is in fact $A$ and $f$ is the morphism associated with $A$.

[^6]:    ${ }^{1}$ Not all morphisms admit an image or coimage in a $k$-linear category since these constructions depend on the existence of kernels and cokernels of certain maps-which cannot be always ensured in this case.
    ${ }^{2}$ Here we are abusing the notation: coker $f$ stands for the morphism associated to the cokernel of $f$, while $\operatorname{ker}(\operatorname{coker} f)$ is the kernel object associated to the morphism coker $f$. The same analogous abuse of notation is used while defining the coimage of $f$.

[^7]:    ${ }^{1}$ This proof is also true for modules over division rings.

[^8]:    ${ }^{1}$ The matrix $T^{*}$ denotes the transpose of $T$, which is the same as the dual of $T$.
    ${ }^{2}$ The matrix $T^{\dagger}$ denotes the complex transpose of $T$.

[^9]:    ${ }^{3}$ In this case, we are using $\operatorname{Mor}_{G r p}(G,-)$ to denote the same as the collection of all group morphisms whose source is $G$.

[^10]:    ${ }^{4}$ As a reminder: $\mathrm{Ab}_{W}$ is the category whose objects are morphisms $f \in \operatorname{Mor}_{\mathrm{Ab}}(-, W)$. If $f: G \rightarrow W$ and $g: H \rightarrow W$ are objects of $\mathrm{Ab}_{W}$, a morphism $h \in \operatorname{Mor}_{\mathrm{Ab}_{W}}(f, g)$ is such that $h g=f$.

[^11]:    ${ }^{1}$ Using the well order on $(m, n)$, let $\ell:=a_{0} m+b_{0} n$ be the smallest element of $(m, n)$. Notice that given any other $u=a m+b n \in(m, n)$, by the euclidean division algorithm there exists two integers $q, r \in \mathbf{Z}$ for which $u=q \ell+r$ and $0 \leqslant r<\ell$. Therefore, one can write

    $$
    r=u-q \ell=(a m+b n)-q\left(a_{0} m+b_{0} n\right)=\left(a-q a_{0}\right) m+\left(b-q b_{0}\right) n
    $$

    so that $r \in(m, n)$ - since $\ell$ is the smallest element, then $r=0$ and thus $\ell$ divides $u$.
    In general, we have shown that $\ell$ divides every element of $(m, n)$, and in particular $\ell$ also divides both $m$ and $n$. Now, if $c \in \mathbf{Z}$ is any common divisor of $m$ and $n$ then in particular $c$ divides $a_{0} m+b_{0} n=\ell-$ showing that $c \leqslant \ell$ and hence $\ell=\operatorname{gcd}(m, n)$.

[^12]:    ${ }^{2} F$ can be covariant or contravariant.

[^13]:    ${ }^{1}$ The annihilator of a module $M$ is defined to be the collection

    $$
    \text { Ann } M:=\{r \in R: r m=0 \text { for all } m \in M\}
    $$

[^14]:    ${ }^{1}$ It is to be noted the subtlety of not adding a dashed arrow (denoting uniqueness) for $\phi$ in the diagram-this is done purposefully since the diagram is commutative in the category of sets, so there may well be a distinct set-function $F_{R} S \rightarrow M$ also making the diagram commute in Set.

[^15]:    ${ }^{2}$ Let $R$ be a commutative ring. If $\mathfrak{a} \subseteq R$ is an ideal, and $M$ is an $R$-module, we there exists a natural structure of $(R / a)$-module that can be endowed on the abelian group $M /(\mathfrak{a} M)$ via the multiplication

    $$
    (r+\mathfrak{a}) \cdot(m+\mathfrak{a} M):=r m+\mathfrak{a} M
    $$

[^16]:    ${ }^{3}$ The collection of pairs $\left(x_{j}, \phi_{j}: P \rightarrow R\right)_{j \in J}$ is sometimes referred to as the dual "basis", but it should be noted right away that such family may not form a basis for the module $P$ —not every projective module is free!

[^17]:    ${ }^{1}$ Where, for any $e \in E, e^{\prime} \in E^{\prime}$, and $(x, y) \in\langle V\rangle^{2}$, we define the maps $s \times t, s^{\prime} \times t^{\prime}$ and $\alpha_{V} \times \alpha_{V}$ as
    $-e \stackrel{s \times t}{\longmapsto}(s(e), t(e))=(t(e), s(e))$.
    $-e^{\prime} \stackrel{s^{\prime} \times t^{\prime}}{\longmapsto}\left(s^{\prime}\left(e^{\prime}\right), t^{\prime}\left(e^{\prime}\right)\right)=\left(t^{\prime}\left(e^{\prime}\right), s^{\prime}\left(e^{\prime}\right)\right)$.
    $-(x, y) \stackrel{\alpha_{V} \times \alpha_{V}}{\longmapsto}\left(\alpha_{V}(x), \alpha_{V}(y)\right)=\left(\alpha_{V}(y), \alpha_{V}(x)\right)$.

[^18]:    ${ }^{2} \amalg$ denotes the standard disjoint union in Set.

[^19]:    ${ }^{1}$ If $z=r e^{\mathrm{i} \theta} \in \mathbf{C}$ is a complex number, in polar form, satisfying $r>0$ and $\theta \in(-\pi, \pi)$, then the complex $\log$ arithm of $z$ is $\log (z)=\log (r)+\mathrm{i} \theta$.

[^20]:    ${ }^{1}$ That is, composed of connected open subsets of $X$.

[^21]:    ${ }^{2}$ Simply refer to the proof of Theorem 16.1.1, now it should be clear that the proof follows exactly the same steps

[^22]:    ${ }^{3}$ It should be noted that the pullback of $p$ over $U$ is nothing more than the induced bundle $\iota^{*} p$ given by the inclusion $t: U \hookrightarrow B$.
    ${ }^{4}$ Here, discrete means a space together with the discrete topology. We shall make use of the functor Disc: Set $\rightarrow$ Top mapping a bare set $S$ to the topological space Disc $S$ with underlying set $S$ and endowed with the discrete topology.

[^23]:    ${ }^{5}$ Since the end-points are fixed—which was pointed out in item (c).

[^24]:    ${ }^{1}$ Given any space $Z$, if $C \subseteq Z$ is a closed set then, given any compact set $K$ and continuous map $g: K \rightarrow Z$, we are always ensured that $g^{-1} C$ is closed-this merely follows from the continuity of $g$. Therefore $k Z$ preserves the closed sets of $Z$, in the sense that if $C$ is closed in $Z$ then $C$ is also closed in $k Z$

[^25]:    ${ }^{2}$ The colimit of the sequence is endowed with the weak topology (see Definition 17.1.21).

[^26]:    ${ }^{1}$ A $k$-frame in an $n$-dimensional vector space is an ordered collection of $k$ linearly independent vectors.

[^27]:    ${ }^{2}$ The use of $\omega_{n}$ denotes the standard orientation for the euclidean space $\mathbf{R}^{n}$ (see Definition 6.6.13).

[^28]:    ${ }^{1} \mathrm{~A}$ simplex with such property is called $\mathcal{U}$-small.

[^29]:    ${ }^{1}$ We denote by tOrd the category of totally ordered sets.

[^30]:    ${ }^{1}$ Empty sequences are also admitted.

[^31]:    ${ }^{1}$ Has all small limits and colimits by Definition 2.5.10.

[^32]:    ${ }^{1}$ That is, there exists a neighbourhood $X \subseteq M$ of $p$ such that the induced map $F: X \rightarrow F(X)$ is a $C^{\infty}$-isomorphism-then $(X, F)$ is a coordinate chart about $p$.

[^33]:    ${ }^{2}$ This is the common description when authors call embedded submanifolds by "regular submanifolds"

[^34]:    ${ }^{1}$ This is due to the fact that 1-manifolds are $C^{\infty}$-isomorphic to a finite disjoint union of circles or intervals (see Corollary 23.8.4).

[^35]:    ${ }^{1}$ Beware! This is not the norm we shall adopt for our studies on banachable topological vector spaces. For the latter, see Definition 26.3.4

[^36]:    ${ }^{1}$ To be precise, $q$ represents the quantile function of the distribution associated with $X$-random variables with equal distribution will have equal quantile functions.

[^37]:    ${ }^{2}$ That is, for each $t \in(0,1)$ and $x, y \in \mathbf{R}^{n}$ we have $g(t x+(1-t) y) \geqslant \operatorname{tg}(x)+(1-t) g(y)$.

[^38]:    ${ }^{3}$ When data is scarce, it may be acceptable to use $S=N$. This is known as the leave-one-out technique for dataset manipulations.
    ${ }^{4}$ The algorithm will define disjoint spaces $R_{k}$ as a mean of classifying future incoming data points $x$ as follows: if $x \in R_{k}$ then predict that $x$ pertains to the class $C_{k}$.

[^39]:    ${ }^{1}$ In reality, this predictor cannot be used since the learner does not have access to the probability distribution $\mathcal{D}$.

[^40]:    ${ }^{2}$ If such points do not exist, we can set $a_{0}:=-\infty$ and $a_{1}:=\infty$.
    ${ }^{3}$ If the first set is empty, we instead set $b_{0}:=\infty$, and if the latter is empty then $b_{1}:=-\infty$.

[^41]:    ${ }^{2}$ If $\sigma=\left[p_{0}, \ldots, p_{k}\right]$, we say that $\mu=\operatorname{sign}(\tau) \sigma$ if $\mu=\left[p_{\tau(0)}, \ldots, p_{\tau(k)}\right]$ - where $\tau \in S_{k+1}$ is a permutation on $k+1$ elements.

