# Mathematics for Natural Scientists II 

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## CHAPTER 1

## Linear spaces and linear maps

In this chapter we study the basic properties of the linear spaces-also called vector spaces-and of the linear maps between them. A linear space is a set endowed with a linear structure, and a linear map between linear spaces is a function between their corresponding sets that preserves their linear structure.

### 1.1. Linear spaces

Definition 1.1.1. A linear space, or a vector space, over $\mathbb{R}$ is a structure $\mathcal{V}:=(X ;+, \mathbf{0}, \cdot)$, where $X$ is a set, $\mathbf{0} \in X$, and,$+ \cdot$ are functions

$$
\begin{array}{r}
+: X \times X \rightarrow X, \quad .: \mathbb{R} \times X \rightarrow X \\
(x, y) \mapsto x+y, \quad(a, x) \mapsto a \cdot x
\end{array}
$$

such that the following conditions are satisfied:

$$
\begin{aligned}
& \left(\mathrm{LS}_{1}\right) \forall_{x, y, z \in X}((x+y)+z=x+(y+z)) . \\
& \left(\mathrm{LS}_{2}\right) \forall_{x \in X}(x+\mathbf{0}=\mathbf{0}+x=x) . \\
& \left(\mathrm{LS}_{3}\right) \forall_{x \in X} \exists_{y \in X}(x+y=\mathbf{0}) . \\
& \left(\mathrm{LS}_{4}\right) \forall_{x, y \in X}(x+y=y+x) . \\
& \left(\mathrm{LS}_{5}\right) \forall_{x, y \in X} \forall_{a \in \mathbb{R}}(a \cdot(x+y)=a \cdot x+a \cdot y) . \\
& \left(\mathrm{LS}_{6}\right) \forall_{x \in X} \forall_{a, b \in \mathbb{R}}((a+b) \cdot x=a \cdot x+b \cdot x) . \\
& \left(\mathrm{LS}_{7}\right) \forall_{x \in X} \forall_{a, b \in \mathbb{R}}((a b) \cdot x=a \cdot(b \cdot x)) . \\
& \left(\mathrm{LS}_{8}\right) \forall_{x \in X}(1 \cdot x=x) .
\end{aligned}
$$

For simplicity, we may write $a x$ instead of $a \cdot x$. The triple $(+, \mathbf{0}, \cdot)$ is called the signature of the linear space $\mathcal{V}$. If, instead of $\mathbb{R}$, we consider any field ${ }^{1} \mathbb{F}$, the

[^0]corresponding structure is called a linear space over $\mathbb{F}$. A linear space over $\mathbb{R}$ is also called a real linear space, and a linear space over the field of complex numbers $\mathbb{C}$ is called a complex linear space. If $\mathcal{V}$ is a linear space, the elements of $X$ are traditionally called vectors. A linear space is called non-trivial, if it contains a vector $x$ such that $x \neq \mathbf{0}$. Unless stated otherwise, the linear spaces considered here are going to be real. When the linear structure on $X$ is clear from the context, we use for simplicity $X$ to denote the vector space $\mathcal{V}$.

Recall that if $X, Y$ are sets, then

$$
X \times Y:=\{(x, y) \mid x \in X \& y \in Y\}
$$

and if $(x, y),\left(x^{\prime}, y^{\prime}\right) \in X \times Y$, then

$$
(x, y)=\left(x^{\prime}, y^{\prime}\right) \Leftrightarrow\left(x=x^{\prime} \& y=y^{\prime} .\right.
$$

Example 1.1.2. Let the structure $\mathcal{R}^{n}:=\left(\mathbb{R}^{n} ;+, \mathbf{0}, \cdot\right)$, where

$$
\begin{gathered}
\mathbb{R}^{n}:=\left\{\left(x_{1}, \ldots, x_{n}\right) \mid x_{1} \in \mathbb{R} \& \ldots \& x_{n} \in \mathbb{R}\right\}, \\
\left(x_{1}, \ldots, x_{n}\right)=\left(y_{1}, \ldots, y_{n}\right) \Leftrightarrow x_{1}=y_{1} \& \ldots \& x_{n}=y_{n}, \\
\left(x_{1}, \ldots, x_{n}\right)+\left(y_{1}, \ldots, y_{n}\right):=\left(x_{1}+y_{1}, \ldots, x_{n}+y_{n}\right), \\
0:=(0, \ldots, 0), \\
a \cdot\left(x_{1}, \ldots, x_{n}\right):=\left(a x_{1}, \ldots, a x_{n}\right) .
\end{gathered}
$$

Clearly, $\mathcal{R}^{n}$ a linear space over $\mathbb{R}$, and, similarly, $\mathcal{Q}^{n}:=\left(\mathbb{Q}^{n} ;+, \mathbf{0}, \cdot\right)$ is linear space over $\mathbb{Q}$, and $\mathcal{C}^{n}:=\left(\mathbb{C}^{n} ;+, \mathbf{0}, \cdot\right)$ is a linear space over $\mathbb{C}$.

If $\mathbb{F}(X, Y)$ is the set of all functions from $X$ to $Y$, and $f, g \in \mathbb{F}(X, Y)$, then

$$
f=g \Leftrightarrow \forall_{x \in X}(f(x)=g(x)) .
$$

Example 1.1.3. If $X$ is a set, $\mathbb{F}(X)$ is the set of all functions $f: X \rightarrow \mathbb{R}$, and if we define the functions $f+g, \overline{0}^{X}$ and $a \cdot f$, where $a \in \mathbb{R}$, by

$$
\begin{aligned}
(f+g)(x) & :=f(x)+g(x), \\
\overline{0}^{X}(x) & :=0, \\
(a \cdot f)(x) & :=a f(x),
\end{aligned}
$$

for every $x \in X$, then $\mathcal{F}(X):=\left(\mathbb{F}(X) ;+\overline{0}^{X}, \cdot\right)$ is a linear space over $\mathbb{R}$.
$\forall_{x \in \mathbb{F}}\left(x \neq \mathbf{0} \Rightarrow \exists_{y \in \mathbb{F}}(x \cdot y=\mathbf{1})\right)$.
It is immediate to see that the rational numbers $\mathbb{Q}$, the real numbers $\mathbb{R}$ and the complex numbers $\mathbb{C}$ have a field structure. Actually, $\mathbb{Q}$ is a subfield of $\mathbb{R}$ and $\mathbb{R}$ is a subfield of $\mathbb{C}$ i.e., the field-signature $(+, \mathbf{0}, \cdot, \mathbf{1})$ of $\mathbb{Q}$ is inherited from the field-signature of $\mathbb{R}$, which, in turn, can be inherited from the field-signature of $\mathbb{C}$.

The Example 1.1.3 shows that a mathematical object can be viewed as a vector, although no immediate geometric intuition is associated with it. If

$$
\boldsymbol{n}:=\{0,1, \ldots, n-1\}
$$

though, an element of $\mathbb{R}^{n}$ can be identified with a function $f: \boldsymbol{n} \rightarrow \mathbb{R}$, and then the Example 1.1.2 is a special case of the Example 1.1.3. If $f, g \in \mathbb{F}(X)$ and $a \in \mathbb{R}$,

$$
\begin{gathered}
f \leq g \Leftrightarrow \forall_{x \in X}(f(x) \leq g(x)) \\
\left.f \leq a: \Leftrightarrow f \leq \bar{a}^{X} \Leftrightarrow \forall_{x \in X}(f(x) \leq a)\right)
\end{gathered}
$$

where $\bar{a}^{X}(x):=a$, for every $x \in X$.
Remark 1.1.4. Let $\mathcal{V}:=(X ;+, \mathbf{0}, \cdot)$ be a linear space, $a, b \in \mathbb{R}$, and $x, y, z, w \in$ $X$. The following hold:
(i) If $z=w$ and $x=y$, then $z+x=w+y$.
(ii) If $x=y$ and $a=b$, then $a \cdot x=b \cdot y$.
(iii) If $x+y=x+z=\mathbf{0}$, then $y=z$.
(iv) $0 \cdot x=\mathbf{0}$.
$(v)(-1) \cdot x=-x$, where, because of case $(i i i),-x$ is the unique element $y$ of $X$ in condition $\left(\mathrm{LS}_{3}\right)$ such that $x+y=\mathbf{0}$.
(vi) If $x \neq \mathbf{0}$ and $a \cdot x=\mathbf{0}$, then $a=0$.

Proof. Exercise.
Definition 1.1.5. Let $\mathcal{V}:=(X ;+, \mathbf{0}, \cdot)$ be a linear space, and $Y \subseteq X$ such that the following conditions are satisfied:
(i) $\forall_{y, y^{\prime} \in Y}\left(y+y^{\prime} \in Y\right)$,
(ii) $\mathbf{0} \in Y$,
(iii) $\forall_{y \in Y} \forall_{a \in \mathbb{R}}(a \cdot y \in Y)$.

Then the structure

$$
\mathcal{V}_{\mid Y}:=\left(Y,+_{\mid Y \times Y}, \mathbf{0}, \cdot_{\mathbb{R} \times Y}\right),
$$

where $+_{\mid Y \times Y}$ is the restriction of + to $Y \times Y$ and ${ }_{\mid \mathbb{R} \times Y}$ is the restrictions of $\cdot$ to $\mathbb{R} \times Y$, is called a linear subspace of $\mathcal{V}$, or, simpler, a subspace of $\mathcal{V}$. We write $\mathcal{V}_{\mid Y} \preceq \mathcal{V}$ to denote that $\mathcal{V}_{\mid Y}$ is a linear subspace of $\mathcal{V}$, although, for simplicity, we refer to a linear subspace $\mathcal{V}_{\mid Y}$ mentioning only the set $Y$, and we write $Y \preceq X$. We denote by $\operatorname{Sub}(\mathcal{V})$ the set of all subspaces of $\mathcal{V}$.

Clearly, $\{0\}$ and $X$ are linear subspaces of $X$.
Example 1.1.6. If $\mathbb{F}^{*}(X)$ is the set of all bounded functions in $\mathbb{F}(X)$ i.e.,

$$
\mathbb{F}^{*}(X)=\left\{f \in \mathbb{F}(X) \mid \exists_{M>0} \forall_{x \in X}(|f(x)| \leq M)\right\},
$$

then $\mathbb{F}^{*}(X)$ is a linear subspace of $\mathbb{F}(X)$ (see Example 1.1.3). To see this let $f, g \in \mathbb{F}(X)$ and $M_{f}>0, M_{g}>0$, such that $|f| \leq M_{f}$ and $|g| \leq M_{g}$. Then
$|f+g| \leq M_{f}+M_{g}$ and $|a f| \leq(1+|a|) M_{f}$, where $M_{f}+M_{g}>0$ and $(1+|a|) M_{f}>0$. Recall that $|f| \in \mathbb{F}(X)$ is defined by $|f|(x):=|f(x)|$, for every $x \in X$.

Example 1.1.7. If $\mathcal{V}:=(X ;+, \mathbf{0}, \cdot)$ is a linear space, $n \geq 1$, and $x_{1}, \ldots, x_{n} \in$ $X$, the set

$$
\left\langle\left\{x_{1}, \ldots, x_{n}\right\}\right\rangle:=\left\{a_{1} \cdot x_{1}+\ldots+a_{n} \cdot x_{n} \mid a_{1} \in \mathbb{R} \& \ldots \& a_{n} \in \mathbb{R}\right\}
$$

is a linear subspace of $\mathcal{V}$. We call an element

$$
\sum_{i=1}^{n} a_{i} x_{i}:=a_{1} \cdot x_{1}+\ldots+a_{n} x_{n}
$$

of $\left\langle\left\{x_{1}, \ldots, x_{n}\right\}\right\rangle$ a linear combination of $x_{1}, \ldots, x_{n}$, and the space $\left\langle\left\{x_{1}, \ldots, x_{n}\right\}\right\rangle$ the linear span of $x_{1}, \ldots, x_{n}$. We may write $\left\langle x_{1}, \ldots, x_{n}\right\rangle$ instead of $\left\langle\left\{x_{1}, \ldots, x_{n}\right\}\right\rangle$.

If $e_{1}:=(1,0), e_{2}:=(0,1),(x, y) \in \mathbb{R}^{2}$, we get $\mathbb{R}^{2}=\left\langle e_{1}, e_{2}\right\rangle$, since

$$
(x . y):=x(1,0)+(0,1) y:=x e_{1}+y e_{2} .
$$

Proposition 1.1.8. Let $\mathcal{V}:=(X ;+, \mathbf{0}, \cdot)$ be a linear space, $Y \subseteq X$, and let $U, V \preceq X$.
(i) If $U+V:=\{u+v \mid u \in U \& v \in V\}$, then $U+V \preceq X$.
(ii) If $U \cap V:=\{x \in X \mid x \in U \& x \in V\}$, then $U \cap V \preceq X$.
(iii) If we define

$$
\langle Y\rangle:=\bigcap\{U \preceq X \mid Y \subseteq U\}:=\left\{x \in X \mid \forall_{U \preceq X}(Y \subseteq U \Rightarrow x \in U)\right\}
$$

then $\langle Y\rangle$ is well-defined (i.e., the set $\{U \preceq X \mid Y \subseteq Y\}$ is non-empty) and it is the least linear subspace of $X$ that includes $Y$.
(iv) If $Y \neq \emptyset$, then

$$
\langle Y\rangle=\left\{\sum_{i=1}^{n} a_{i} y_{i} \mid n \geq 1 \& \forall_{i \in\{1, \ldots, n\}}\left(a_{i} \in \mathbb{R} \& y_{i} \in Y\right)\right\}
$$

Proof. Exercise.
Since $\emptyset \subseteq\{\mathbf{0}\}$, we have that $\langle\emptyset\rangle=\{\mathbf{0}\}$. The subspace $U+V$ of $\mathcal{X}$ is called the sum of $U$ and $V$. By Proposition 1.1.8 the linear span $\left\langle x_{1}, \ldots, x_{n}\right\rangle$ of $x_{1}, \ldots, x_{n} \in X$ is the least linear space containing $x_{1}, \ldots, x_{n}$. If $X=\langle Y\rangle$, we say that $Y$ generates the linear space $\mathcal{V}$ (or $X)$, and the elements of $Y$ are called generators of $\mathcal{V}$.

### 1.2. Finite-dimensional linear spaces

Definition 1.2.1. Let $\mathcal{V}:=(X ;+, \mathbf{0}, \cdot)$ be a linear space, $n \geq 1$, and let $x_{1}, \ldots, x_{n} \in X$. We say that the vectors $x_{1}, \ldots, x_{n}$ are linearly dependent, or that
their set $\left\{y_{1}, \ldots, y_{n}\right\}$ is a linearly dependent subset of $X$, if

$$
\exists_{a_{1}, \ldots, a_{n} \in \mathbb{R}}\left(\exists_{i \in\{1, \ldots, n\}}\left(a_{i} \neq 0\right) \& \sum_{i=1}^{n} a_{i} x_{i}=\mathbf{0}\right) .
$$

We say that $x_{1}, \ldots, x_{n}$ are linearly independent, if they are not linearly dependent. A subset $Y$ of $X$ is called linearly dependent, if

$$
\exists_{n \geq 1} \exists_{y_{1}, \ldots, y_{n} \in Y}\left(\left\{y_{1}, \ldots, y_{n}\right\} \text { is linearly dependent }\right)
$$

while it is called linearly independent, if it is not a linearly dependent subset of $X$.
If $x_{1}, \ldots, x_{n}$ are linearly dependent, $a_{1} x_{1}+\ldots+a_{n} x_{n}=\mathbf{0}$, and $a_{i} \neq 0$, then

$$
x_{i}=\left(\frac{-a_{1}}{a_{i}}\right) x_{1}+\ldots+\left(\frac{-a_{i-1}}{a_{i}}\right) x_{i-1}+\left(\frac{-a_{i+1}}{a_{i}}\right) x_{i+1}+\ldots+\left(\frac{-a_{n}}{a_{i}}\right) x_{n}
$$

i.e., $x_{i}$ is a linear combination of $a_{1}, \ldots, a_{i-1}, a_{i+1}, \ldots, a_{n}$.

Remark 1.2.2. Let $X$ be a linear space and $Y, Z \subseteq X$.
(i) If $x_{1}, \ldots, x_{n} \in X$, then $x_{1}, \ldots, x_{n}$ are linearly independent if and only if

$$
\forall_{a_{1}, \ldots, a_{n} \in \mathbb{R}}\left(\sum_{i=1}^{n} a_{i} x_{i}=\mathbf{0} \Rightarrow \forall_{i \in\{1, \ldots, n\}}\left(a_{i}=0\right)\right)
$$

(ii) $Y$ is linearly independent if and only if

$$
\forall_{n \geq 1} \forall_{y_{1}, \ldots, y_{n} \in Y}\left(\left\{y_{1}, \ldots, y_{n}\right\} \text { is linearly independent }\right) .
$$

(iii) $\{\mathbf{0}\}$ and $X$ are linearly dependent subsets of $X$.
(iv) If $x \neq \mathbf{0}$, then $\{x\}$ is a linearly independent subset of $X$.
(v) The empty set $\emptyset$ is a linearly independent subset of $X$.
(vi) If $Y$ is linearly dependent and $Y \subseteq Z$, then $Z$ is linearly dependent.
(vii) If $Y$ is linearly independent and $Z \subseteq Y$, then $Z$ is linearly independent.

Proof. (i) and (ii) By negating the corresponding defining formulas.
(iii) $1 \cdot \mathbf{0}=\mathbf{0}$, and $\{\mathbf{0}\}$ is a linearly dependent subset of $X$.
(iv) It follows immediately by Remark 1.1.4(vi).
(v) If we suppose that $\emptyset$ is a linearly dependent subset of $X$ i.e.,

$$
\exists_{n \geq 1} \exists_{y_{1}, \ldots, y_{n}}\left(y_{1} \in \emptyset \& \ldots \& y_{n} \in \emptyset \&\left\{y_{1}, \ldots, y_{n}\right\} \text { is linearly dependent }\right)
$$

it is immediate that we get a contradiction from it.
(vi) and (vii) are immediate to show.

Example 1.2.3. The following $n$-vectors in $\mathbb{R}^{n}$

$$
e_{1}:=(1,0, \ldots, 0), \quad e_{2}:=(0,1,0, \ldots, 0), \ldots, e_{n}:=(0, \ldots, 0,1)
$$

are linearly independent, since for every $a_{1}, \ldots, a_{n} \in \mathbb{R}$ we have that

$$
\sum_{i=1}^{n} a_{i} e_{i}=0 \Leftrightarrow\left(a_{1}, \ldots, a_{n}\right)=\mathbf{0} \Leftrightarrow a_{1}=\ldots=a_{n}=0
$$

Example 1.2.4. For every $n \geq 1$, the following $n$-vectors in $\mathbb{F}(\mathbb{R})$

$$
f_{1}(t):=e^{t}, \ldots, f_{n}(t):=e^{n t}
$$

are linearly independent (Exercise).
Remark 1.2.5. Let $\mathcal{V}:=(X ;+, \mathbf{0}, \cdot)$ be a linear space, $n \geq 1$, and $x_{1}, \ldots, x_{n} \in$ $X$ linearly independent. If $a_{1}, \ldots, a_{n}, b_{1}, \ldots, b_{n} \in \mathbb{R}$, then

$$
\sum_{i=1}^{n} a_{i} x_{i}=\sum_{i=1}^{n} b_{i} x_{i} \Rightarrow\left(a_{1}=b_{1} \& \ldots \& a_{n}=b_{n}\right)
$$

Moreover, $x_{i} \neq \mathbf{0}$, for every $i \in\{1, \ldots, n\}$.
Proof. It follows from the Definition 1.2.1 and the equivalence

$$
\sum_{i=1}^{n} a_{i} x_{i}=\sum_{i=1}^{n} b_{i} x_{i} \Leftrightarrow \sum_{i=1}^{n}\left(a_{i}-b_{i}\right) x_{i}=\mathbf{0}
$$

If there is $i \in\{1, \ldots, n\}$ such that $x_{i}=0$, then $0 x_{1}+0 x_{i-1}+1 x_{i}+0 x_{i+1}+\ldots+$ $0 x_{n}=\mathbf{0}$, which is impossible.

Definition 1.2.6. If $\mathcal{V}:=(X ;+, \mathbf{0}, \cdot)$ is linear space, a subset $B$ of $X$ is called a basis of $\mathcal{V}$ (or, for simplicity a basis of $X$ ), if $B$ is linearly independent, and $\langle B\rangle=X$. If $\mathcal{V}$ has a finite basis $B$, it is called a finite-dimensional linear space, while if it has an infinite basis, it is called infinite-dimensional.

Consequently, the subspace $\{\mathbf{0}\}$ has as a basis the empty set.
Example 1.2.7. The set $E_{n}:=\left\{e_{1}, \ldots, e_{n}\right\}$ of the linearly independent elements in $\mathbb{R}^{n}$ that were defined in the Example 1.2 .3 is the standard basis of $\mathbb{R}^{n}$. Hence, $\mathcal{R}^{n}$ is finite-dimensional. It is easy to see that $\mathbb{R}^{n}$ has more than one bases. E.g., $B:=\{(1,1),(-1,2)\}$ is another basis of $\mathbb{R}^{2}$.

Example 1.2.8. Since the set $E:=\left\{e^{n t} \mid n \geq 1\right\}$ is a linearly independent subset of $\mathbb{F}(\mathbb{R})$, the set $E$ is a basis of the linear subspace $\langle E\rangle$ of $\mathbb{F}(\mathbb{R})$, and $\langle E\rangle$ is infinite-dimensional.

Corollary 1.2.9. Let $\mathcal{V}:=(X ;+, \mathbf{0}, \cdot)$ be a linear space, and $x \in X$. If $B:=\left\{v_{1}, \ldots, v_{n}\right\}$ is a basis of $\mathcal{V}$, there are unique $a_{1}, \ldots, a_{n} \in \mathbb{R}$ such that

$$
x=\sum_{i=1}^{n} a_{i} v_{i} .
$$

Proof. It follows by the definition of a basis and the Remark 1.2.5.
These unique $a_{1}, \ldots, a_{n} \in \mathbb{R}$ are called the coordinates of $x$ with respect to $B$.
Definition 1.2.10. Let $\mathcal{V}:=(X ;+, \mathbf{0}, \cdot)$ be a linear space, $\left\{v_{1}, \ldots, v_{n}\right\} \subseteq$ $X$ and $m \leq n$. The set $\left\{v_{1}, \ldots v_{m}\right\}$ is a maximal subset of linearly independent elements of $X$, if it is a linearly independent subset of $X$, and for every $k \in \mathbb{N}$, such that $m<k \leq n$, the set $\left\{v_{1}, \ldots, v_{m}, v_{k}\right\}$ is a linearly dependent subset of $X$.

Theorem 1.2.11 (Finite basis-criterion I). Let $\mathcal{V}:=(X ;+, \mathbf{0}, \cdot)$ be a linear space, $n \geq 1$, and $\left\{v_{1}, \ldots, v_{n}\right\} \subseteq X$ such that $X=\left\langle\left\{v_{1}, \ldots, v_{n}\right\}\right\rangle$. If $\left\{v_{1}, \ldots, v_{r}\right\}$ is a maximal subset of linearly independent elements of $X$, where $1 \leq r \leq n$, then $\left\{v_{1}, \ldots, v_{r}\right\}$ is a basis of $\mathcal{V}$.

Proof. If $r=n$, then $\left\{v_{1}, \ldots, v_{r}\right\}$ is a linearly independent subset generating $X$ i.e., it is a basis of $\mathcal{V}$. If $r<n$, by the maximality of $\left\{v_{1}, \ldots, v_{r}\right\}$ the sets

$$
\left\{v_{1}, \ldots, v_{r}, v_{r+1}\right\},\left\{v_{1}, \ldots, v_{r}, v_{r+2}\right\}, \ldots,\left\{v_{1}, \ldots, v_{r}, v_{n}\right\}
$$

are linearly dependent subsets of $X$. We show that

$$
v_{r+1} \in\left\langle\left\{v_{1}, \ldots, v_{r}\right\}\right\rangle \& v_{r+2} \in\left\langle\left\{v_{1}, \ldots, v_{r}\right\}\right\rangle \& \ldots \& v_{n} \in\left\langle\left\{v_{1}, \ldots, v_{r}\right\}\right\rangle
$$

We show this only for $v_{r+1}$, and for $v_{r+2}, \ldots, v_{n}$ we proceed similarly. Since $\left\{v_{1}, \ldots, v_{n}, v_{r+1}\right\}$ is linearly dependent, there are $a_{1}, \ldots, a_{r}, a_{r+1} \in \mathbb{R}$ such that

$$
a_{1} v_{1}+\ldots+a_{r} v_{r}+a_{r+1} v_{r+1}=\mathbf{0}
$$

and not all of them are equal to 0 . If $a_{r+1}=0$, then $a_{1} v_{1}+\ldots+a_{r} v_{r}=\mathbf{0}$, hence $a_{1}=\ldots=a_{r}=a_{r+1}=0$, which is a contradiction. Hence $a_{r+1} \neq 0$, and hence $v_{r+1}$ can be written as a linear combination of $v_{1}, \ldots, v_{r}$. Since an element $x$ of $X$ is a linear combination of $v_{1}, \ldots, v_{r}, v_{r+1}, \ldots, v_{n}$ and $v_{r+1}, \ldots, v_{n}$ are linear combinations of $v_{1}, \ldots, v_{r}$, then $x$ is a linear combination of $v_{1}, \ldots, v_{r}$.

Next we show that we can replace any number of elements of a finite basis by an equal number of any linearly independent vectors.

Lemma 1.2.12 (Exchange lemma (Steinitz)). Let $n, m \geq 1,\left\{v_{1}, \ldots v_{n}\right\}$ a basis of the linear space $\mathcal{V}:=(X ;+, \mathbf{0}, \cdot)$, and let $w_{1}, \ldots, w_{m} \in X$ be linearly independent.
(i) If $m<n$, there are $u_{m+1}, \ldots, u_{n} \in\left\{v_{1}, \ldots v_{n}\right\}$ such that

$$
\left\langle\left\{w_{1}, \ldots, w_{m}, u_{m+1}, \ldots, u_{n}\right\}\right\rangle=X
$$

(ii) If $m=n$, then $\left\langle\left\{w_{1}, \ldots, w_{n}\right\}\right\rangle=X$.

Proof. (i) By the definition of a basis there are $a_{1}, \ldots a_{n} \in \mathbb{R}$ such that

$$
w_{1}=a_{1} v_{1}+\ldots+a_{n} v_{n}
$$

Since by Remark 1.2.5 $w_{1} \neq \mathbf{0}$, there is some $a_{i} \neq 0$, where $i \in\{1, \ldots, n\}$. Without loss of generality we can take $i=1$ (if $a_{1}=0$, we can re-enumerate the elements
of the set $\left\{v_{1}, \ldots v_{n}\right\}$ so that the first coefficient in the writing of $w_{1}$ as a linear combination of the elements of the set $\left\{v_{1}, \ldots v_{n}\right\}$ is non-zero). Hence

$$
a_{1} v_{1}=w_{1}-\sum_{i=2}^{n} a_{i} v_{i} \Leftrightarrow v_{1}=\frac{1}{a_{1}} w_{1}-\sum_{i=2}^{n} \frac{a_{i}}{a_{1}} v_{i}
$$

and consequently

$$
v_{1} \in\left\langle\left\{w_{1}, v_{2}, \ldots, v_{n}\right\}\right\rangle
$$

and

$$
\left\langle\left\{w_{1}, v_{2}, \ldots, v_{n}\right\}\right\rangle=X
$$

By the inductive hypothesis, if $1 \leq r<m$ we get (possibly after a re-enumeration of the set $\left\{v_{1}, \ldots v_{n}\right\}$ )

$$
\left\langle\left\{w_{1}, \ldots, w_{r}, v_{r+1}, \ldots, v_{n}\right\}\right\rangle=X
$$

Hence,

$$
w_{r+1}=b_{1} w_{1}+\ldots+b_{r} w_{r}+c_{r+1} v_{r+1}+\ldots+c_{n} v_{n}
$$

Not all the terms $c_{r+1}, \ldots, c_{n}$ are equal to 0 , since then $w_{r+1}$ would be a linear combination of $w_{1}, \ldots, w_{r}$, something that contradicts the hypothesis of linear independence of the vectors $w_{1}, \ldots, w_{m}$. Without loss of generality, let $c_{r+1} \neq 0$, hence

$$
\begin{aligned}
& c_{r+1} v_{r+1}=w_{r+1}-\left[\sum_{i=1}^{r} b_{i} w_{i}+\sum_{j=r+2}^{n} c_{j} v_{j}\right] \Leftrightarrow \\
& v_{r+1}=\frac{1}{c_{r+1}} w_{r+1}-\sum_{i=1}^{r} \frac{b_{i}}{c_{r+1}} w_{i}-\sum_{j=r+2}^{n} \frac{c_{j}}{c_{r+1}} v_{j}
\end{aligned}
$$

and consequently

$$
v_{r+1} \in\left\langle\left\{w_{1}, \ldots, w_{r}, w_{r+1}, v_{r+2}, \ldots, v_{n}\right\}\right\rangle,
$$

and

$$
\left\langle\left\{w_{1}, \ldots, w_{r}, w_{r+1}, v_{r+2}, \ldots, v_{n}\right\}\right\rangle=X
$$

After $m$-number of steps, we get $\left\langle\left\{w_{1}, \ldots, w_{m}, u_{m+1}, \ldots, u_{n}\right\}\right\rangle=X$.
(ii) It follows immediately by (i).

Theorem 1.2.13. Let $0<n<m$, and let $\left\{v_{1}, \ldots v_{n}\right\}$ be a basis of the linear space $\mathcal{V}:=(X ;+, \mathbf{0}, \cdot)$. If $w_{1}, \ldots, w_{m} \in X$, then $w_{1}, \ldots, w_{m}$ are linearly dependent.

Proof. Suppose that the vectors $w_{1}, \ldots, w_{m}$ are linearly independent. Since then the vectors $w_{1}, \ldots, w_{n}$ are also linearly independent, by the Lemma 1.2.12(ii) we have that $w_{1}, \ldots, w_{n}$ is a basis of $X$. By the hypothesis of linear independence we have that $w_{n+1} \neq \mathbf{0}$, hence it is also a non-trivial linear combination of $w_{1}, \ldots, w_{n}$. By this contradiction we conclude that the vectors $w_{1}, \ldots, w_{m}$ are linearly dependent.

Corollary 1.2.14. If $B_{1}, B_{2}$ are finite bases of a linear space $\mathcal{V}$, then $B_{1}$ and $B_{2}$ have the same number of elements.

Proof. If $\mathcal{V}$ is a trivial linear space, then the two bases are equal to the empty set, and $\left|B_{1}\right|=\left|B_{2}\right|=0$, where $|I|$ denotes the number of elements, or the cardinality, of a set $I$. Let $\mathcal{V}$ be non-trivial, and let $n, m \geq 1$ such that $\left|B_{1}\right|=n$ and $\left|B_{2}\right|=m$. If $n<m$, then by the Theorem 1.2 .13 we have that $B_{2}$ is linearly dependent, which is a contradiction. Hence $n \geq m$. Similarly we get $m \geq n$.

Because of the Corollary 1.2.14 the following concept is well-defined.
Definition 1.2.15. If $n \geq 1$ and $\left\{v_{1}, \ldots, v_{n}\right\}$ is a basis of a linear space $\mathcal{V}:=$ $(X ;+, \mathbf{0}, \cdot)$, we call $\mathcal{V}$ an $n$-dimensional space, and we write $\operatorname{dim}(X):=n$. A trivial linear space has dimension 0 .

Clearly, $\operatorname{dim}\left(\mathbb{R}^{n}\right):=n$.
Corollary 1.2.16. Let $n \geq 1$, and let $v_{1}, \ldots, v_{n}$ be linearly independent elements of a linear space $X$.
(i) (Finite basis-criterion II) If their set $M:=\left\{v_{1}, \ldots, v_{n}\right\}$ is a maximal set of linearly independent elements of $X$ i.e., for every $x \in X$ we have that

$$
x, v_{1}, \ldots, v_{n}
$$

are linearly dependent elements of $X$, then $M$ is a basis of $X$.
(ii) If $\operatorname{dim}(X)=n$, and $w_{1}, \ldots, w_{n}$ are linearly independent elements of $X$, then $B:=\left\{w_{1}, \ldots, w_{n}\right\}$ is a basis of $X$.
(iii) If $Y$ is a subspace of $X$ with $\operatorname{dim}(Y)=\operatorname{dim}(X)=n$, then $Y=X$.
(iv) If $\operatorname{dim}(X)=n, 1 \leq r<n$, and $w_{1}, \ldots, w_{r}$ are linearly independent elements of $X$, then there are elements $v_{r+1}, \ldots, v_{n}$ of $X$ such that the set

$$
\left\{w_{1}, \ldots, w_{r}, v_{r+1}, \ldots, v_{n}\right\}
$$

is a basis of $X$.
Proof. Exercise.
Next we show that the existence of a basis of a linear space $X$ implies the existence of a basis of any subspace of $X$.

Corollary 1.2.17. Let $\mathcal{V}:=(X ;+, \mathbf{0}, \cdot)$ be a linear space with $\operatorname{dim}(X)=n$. If $Y \preceq X$, then $Y$ has a basis and $\operatorname{dim}(Y) \leq \operatorname{dim}(X)$.

Proof. If $Y:=\{\boldsymbol{0}\}$, then $\emptyset$ is a basis of $Y$ and $\operatorname{dim}(Y)=0 \leq \operatorname{dim}(X)$. If $Y$ is non-trivial, then either $Y=X$, or $Y$ is a proper subspace of $X$. In the first case what we want to show follows trivially. If $Y$ is a proper, non-trivial subspace of $X$, then there is some $y_{1} \in Y$ such that $y_{1} \neq \mathbf{0}$, and by the Remark 1.1.4(vi) $M_{1}:=\left\{y_{1}\right\}$ is linearly independent. By the principle of the excluded middle ${ }^{2}$

[^1](PEM), we have that $M_{1}$ is either a maximal set of linearly dependent elements of $Y$, hence by the Corollary 1.2.16(i) it is also a basis of $Y$, and hence $\operatorname{dim}(Y)=1$, or there is $y_{2} \in Y$ such that $M_{2}:=\left\{y_{1}, y_{2}\right\}$ is linearly independent. Proceeding similarly, we can repeat the same argument at most $(n-1)$ number of times, in order to reach the required conclusion.

Next we write the expression that abbreviates the unique existence of an element of a set $X$ satisfying a formula $\phi(x)$ :

$$
\exists_{!x \in X}(\phi(x)): \Leftrightarrow \exists_{x \in X}\left(\phi(x) \& \forall_{y \in X}(\phi(y) \Rightarrow y=x)\right) .
$$

Proposition 1.2.18. If $X$ is a linear space, and $Y, Z \preceq X$, such that

$$
\forall_{x \in X} \exists!y \in Y \exists!z \in Z ~(x=y+z),
$$

we write $X:=Y \oplus Z$. The following are equivalent:
(i) $X=Y \oplus Z$.
(ii) $X=Y+Z$ and $Y \cap Z=\{\mathbf{0}\}$.

Proof. Exercise.
Proposition 1.2.19. Let $X$ be a linear space, $n \in \mathbb{N}$, and $\operatorname{dim}(X)=n$.
(i) If $Y \preceq X$, there is some $Z \preceq X$ such that $X=Y \oplus Z$.
(ii) If $Y, Z \preceq X$ such that $X=Y \oplus Z$, then $\operatorname{dim}(X)=\operatorname{dim}(Y)+\operatorname{dim}(Z)$.

Proof. Exercise.
Next we give a condition under which, a linearly independent subset of a linear space $X$ can be extended to a larger linearly independent subset of $X$.

Lemma 1.2.20. Let $Y$ be a linearly independent subset of a linear space $X$, and $x_{0} \in X$. If $x_{0} \notin\langle Y\rangle$, then $Y \cup\left\{x_{0}\right\}$ is a linearly independent subset of $X$.

Proof. Exercise.

### 1.3. Linear maps

Definition 1.3.1. If $X$ and $Y$ are linear spaces, a function $f: X \rightarrow Y$ is called linear, or a linear map, if it satisfies the following conditions:
(i) $\forall_{x, x^{\prime} \in X}\left(f\left(x+x^{\prime}\right)=f(x)+f\left(x^{\prime}\right)\right)$.
(ii) $\forall_{x \in X} \forall_{a \in \mathbb{R}}(f(a \cdot x)=a \cdot f(x))$.

Moreover, we define the following sets:

$$
\begin{aligned}
& \mathcal{L}(X, Y):=\{f: X \rightarrow Y \mid f \text { is linear }\} \\
& \mathcal{L}(X):=\mathcal{L}(X, X):=\{f: X \rightarrow X \mid f \text { is linear }\}
\end{aligned}
$$

$$
X^{*}:=\mathcal{L}(X, \mathbb{R}):=\{f: X \rightarrow \mathbb{R} \mid f \text { is linear }\} .
$$

The elements of $\mathcal{L}(X)$ are called operators on $X$, or linear transformations on $X$, while $X^{*}$ is called the dual space of $X$.

Example 1.3.2. If $X$ is a linear space with $\operatorname{dim}(X)=n$, for some $n \geq 1$, and $B:=\left\{v_{1}, \ldots, v_{n}\right\}$ is a fixed basis of $X$, then the function $f_{B}: X \rightarrow \mathbb{R}^{n}$, defined by

$$
f_{B}(x):=\left(a_{1}, \ldots, a_{n}\right), \quad x=\sum_{i=1}^{n} a_{i} v_{i}
$$

is a linear map. Moreover, if $i \in\{1, \ldots, n\}$, the function $\operatorname{pr}_{i}^{B}: X \rightarrow \mathbb{R}$, defined by

$$
\begin{gathered}
\operatorname{pr}_{i}^{B}(x):=a_{i}, \quad x=\sum_{i=1}^{n} a_{i} v_{i}, \\
\left.X \xrightarrow{f_{B}} \underset{\substack{\mathbb{R}^{n}}}{\operatorname{pr}_{i}^{B}}\right|_{\substack{n}} ^{\operatorname{pr}_{i}}
\end{gathered}
$$

is a linear map. If $n>m \geq 1$, the function $g: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ is linear, where

$$
g\left(a_{1}, \ldots, a_{m}, a_{m+1}, \ldots, a_{n}\right):=\left(a_{1}, \ldots, a_{m}\right)
$$

Remark 1.3.3. The set $\mathcal{L}(X, Y)$ is equipped with the following linear structure

$$
\begin{aligned}
(f+g)(x) & :=f(x)+g(x), \quad x \in X, \\
(a \cdot f)(x) & :=a \cdot f(x), \quad a \in \mathbb{R}, \quad x \in X, \\
\mathbf{0}(x) & :=\mathbf{0}, \quad x \in X .
\end{aligned}
$$

Proof. Exercise.
Definition 1.3.4. If $m, n \geq 1$, an array of real numbers

$$
A:=\left[\begin{array}{ccc}
a_{11} & \ldots & a_{1 n} \\
\vdots & \vdots & \vdots \\
a_{i 1} & \ldots & a_{i n} \\
\vdots & \vdots & \vdots \\
a_{m 1} & \ldots & a_{m n}
\end{array}\right]=:\left[a_{i j}\right] .
$$

is called a matrix of $m$-rows and $n$-columns. If $1 \leq i \leq m$, the $i$-th row of $A$ is the array

$$
A_{i}:=\left[\begin{array}{lll}
a_{i 1} & \ldots & a_{i n}
\end{array}\right]:=\left[a_{i j}\right]_{i},
$$

and if $1 \leq j \leq n$, the $j$-th column of $A$ is the array

$$
A^{j}:=\left[\begin{array}{c}
a_{1 j} \\
\vdots \\
a_{m j}
\end{array}\right]:=\left[a_{i j}\right]^{j}
$$

The set of $m \times n$-matrices is denoted by $M_{m, n}(\mathbb{R})$, while the set of square matrices $M_{n, n}(\mathbb{R})$ is also denoted by $M_{n}(\mathbb{R})$. If $\left[a_{i j}\right],\left[b_{i j}\right] \in M_{m, n}(\mathbb{R})$, and $a \in \mathbb{R}$, we define

$$
\begin{aligned}
& {\left[a_{i j}\right]=\left[b_{i j}\right]: \Leftrightarrow \forall_{i \in\{1, \ldots, m\}} \forall_{j \in\{1, \ldots, n\}}\left(a_{i j}=b_{i j}\right) } \\
& {\left[a_{i j}\right]+\left[b_{i j}\right] }:=\left[a_{i j}+b_{i j}\right], \\
& a \cdot\left[b_{i j}\right]:=\left[a b_{i j}\right], \\
& \mathbf{0}_{m n}:=[0],
\end{aligned}
$$

and if $m=n$, we denote $\mathbf{0}_{n n}$ by $\mathbf{0}_{n}$, or, if $n$ is clear from the context, by $\mathbf{0}$.
If $m=n=2$, the above definitions take the form

$$
\begin{gathered}
{\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right]=\left[\begin{array}{ll}
a^{\prime} & b^{\prime} \\
c^{\prime} & d^{\prime}
\end{array}\right] \Leftrightarrow a=a^{\prime} \& b=b^{\prime} \& c=c^{\prime} \& d=d^{\prime},} \\
{\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right]+\left[\begin{array}{ll}
a^{\prime} & b^{\prime} \\
c^{\prime} & d^{\prime}
\end{array}\right]=\left[\begin{array}{ll}
a+a^{\prime} & b+b^{\prime} \\
c+c^{\prime} & d+d^{\prime}
\end{array}\right],} \\
\lambda\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right]=\left[\begin{array}{ll}
\lambda a & \lambda b \\
\lambda c & \lambda d
\end{array}\right], \quad \lambda \in \mathbb{R}, \\
\mathbf{0}_{2}:=\left[\begin{array}{ll}
0 & 0 \\
0 & 0
\end{array}\right] .
\end{gathered}
$$

It is easy to see that $M_{m, n}(\mathbb{R})$, and as a special case $M_{2}(\mathbb{R})$, equipped with the above operations, is a linear space.

Example 1.3.5. If

$$
A=\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right]
$$

let $f_{A}: \mathbb{R}^{2} \rightarrow \mathbb{R}^{2}$ be defined by

$$
f_{A}(x, y):=A\left[\begin{array}{l}
x \\
y
\end{array}\right]:=\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right]\left[\begin{array}{l}
x \\
y
\end{array}\right]:=\left[\begin{array}{l}
a x+b y \\
c x+d y
\end{array}\right] .
$$

Since

$$
\begin{aligned}
f_{A}\left((x, y)+\left(x^{\prime}, y^{\prime}\right)\right) & :=\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right]\left[\begin{array}{l}
x+x^{\prime} \\
y+y^{\prime}
\end{array}\right] \\
& =\left[\begin{array}{l}
a\left(x+x^{\prime}\right)+b\left(y+y^{\prime}\right) \\
c\left(x+x^{\prime}\right)+d\left(y+y^{\prime}\right)
\end{array}\right] \\
& =\left[\begin{array}{l}
a x+b y \\
c x+d y
\end{array}\right]+\left[\begin{array}{l}
a x^{\prime}+b y^{\prime} \\
c x^{\prime}+d y^{\prime}
\end{array}\right]
\end{aligned}
$$

$$
\begin{aligned}
& =A\left[\begin{array}{l}
x \\
y
\end{array}\right]+A\left[\begin{array}{l}
x^{\prime} \\
y^{\prime}
\end{array}\right] \\
& =f_{A}((x, y))+f_{A}\left(\left(x^{\prime}, y^{\prime}\right)\right)
\end{aligned}
$$

Similarly we show that $f_{A}(\lambda(x, y))=\lambda f_{A}((x, y))$, for every $\lambda \in \mathbb{R}$.
Remark 1.3.6. Let $X, Y, Z$ be linear spaces, $f \in \mathcal{L}(X, Y)$ and $g \in \mathcal{L}(Y, Z)$.
(i) The composite function $g \circ f$ is in $\mathcal{L}(X, Z)$, where $g \circ f: X \rightarrow Z$ is defined by

$$
(g \circ f)(x):=g(f(x)), \quad x \in X
$$

(ii) $\operatorname{id}_{X} \in \mathcal{L}(X)$.
(iii) $f(\mathbf{0})=\mathbf{0}$.
(iv) if $x \in X$, then $f(-x)=-f(x)$.
(v) If $n \geq 1, a_{1}, \ldots a_{n} \in \mathbb{R}$, and $x_{1}, \ldots x_{n} \in X$, then

$$
f\left(\sum_{i=1}^{n} a_{i} x_{i}\right)=\sum_{i=1}^{n} a_{i} f\left(x_{i}\right) .
$$

Proof. Exercise. For the inductive proof of the case (vi), use the following recursive definition of $\sum_{i=1}^{n} x_{i}$, where $x_{1}, \ldots, x_{n} \in X$ and $n \geq 1$ :

$$
\sum_{i=1}^{n} x_{i}:=\left\{\begin{array}{cl}
x_{1} & , n=1 \\
\left(\sum_{i=1}^{n-1} x_{i}\right)+x_{n} & , n>1
\end{array}\right.
$$

A linear map preserves linear dependence, but not necessarily linear independence. The latter holds if a linear map is injective. If it is a bijection i.e., an injection and a surjection, it sends a basis of its domain to a basis of its codomain.

Proposition 1.3.7. If $X, Z$ are linear spaces, $Y \subseteq X, f \in \mathcal{L}(X, Z)$, and $x_{1}, \ldots, x_{n} \in X$, the following hold.
(i) If $x_{1}, \ldots x_{n}$ are linearly dependent in $X$, then $f\left(x_{1}\right), \ldots, f\left(x_{n}\right)$ are linearly dependent in $Z$.
(ii) If $Y$ is a linearly dependent subset of $X$, then $f(Y):=\{f(y) \mid y \in Y\}$ is a linearly dependent subset of $Z$.
(iii) If $x_{1}, \ldots x_{n}$ are linearly independent in $X$, then there is a linear map $g: X \rightarrow Z$ such that $g\left(x_{1}\right), \ldots, g\left(x_{n}\right)$ are linearly dependent in $Z$.
(iv) If $x_{1}, \ldots x_{n}$ are linearly independent in $X$, and if $f$ is an injection, then $f\left(x_{1}\right), \ldots, f\left(x_{n}\right)$ are linearly independent in $Z$.
$(v)$ If $Y$ is a linearly independent subset of $X$, and if $f$ is an injection, then $f(Y)$ is a linearly independent subset of $Z$.
(vi) If $X=\langle Y\rangle$, and if $f$ is a surjection, then $Z=\langle f(Y)\rangle$.
(vii) If $Y$ is a basis of $X$, and if $f$ is a bijection, then $f(Y)$ is a basis of $Z$.

Proof. (i) Let $a_{1}, \ldots a_{n} \in \mathbb{R}$, where $a_{i} \neq 0$, for some $i \in\{1, \ldots, n\}$ such that $\sum_{i=1}^{n} a_{i} x_{i}=\mathbf{0}$. Then what we want follows from the equalities

$$
\mathbf{0}=f(\mathbf{0})=f\left(\sum_{i=1}^{n} a_{i} x_{i}\right)=\sum_{i=1}^{n} a_{i} f\left(x_{i}\right) .
$$

(ii) It follows immediately from the case (i).
(iii) For example, we can take $g$ to be the zero map.
(iv) By the injectivity of $f$, if $a_{1}, \ldots, a_{n} \in \mathbb{R}$, we have that

$$
\begin{aligned}
\sum_{i=1}^{n} a_{i} f\left(x_{i}\right)=\mathbf{0} & \Leftrightarrow f\left(\sum_{i=1}^{n} a_{i} x_{i}\right)=f(\mathbf{0}) \\
& \Leftrightarrow \sum_{i=1}^{n} a_{i} x_{i}=\mathbf{0} \\
& \Leftrightarrow a_{1}=\ldots=a_{n}=0
\end{aligned}
$$

(v) It follows immediately from the case (iv).
(vi) If $X$ is trivial, then $Y=\emptyset$ or $Y=X$. In both cases what we want follows immediately. Let $X$ be non-trivial, and let $z \in Z$. Then there is $x \in X$ such that $f(x)=z$. If $a_{1}, \ldots, a_{n} \in \mathbb{R}$ and $y_{1}, \ldots, y_{n} \in Y$ such that $x=\sum_{i=1}^{n} a_{i} y_{i}$, then

$$
z=f(x)=f\left(\sum_{i=1}^{n} a_{i} y_{i}\right)=\sum_{i=1}^{n} a_{i} f\left(y_{i}\right) \in\langle f(Y)\rangle
$$

(vii) By the case (v) we have that $f(Y)$ is a linearly independent subset of $Z$, and by the case (vi) we have that $Z=\langle f(Y)\rangle$.

A linear map $f: X \rightarrow Y$, which is is a linear isomorphism guarantees that the two linear spaces $X$ and $Y$ are the "same" from the linear-structure point of view.

Definition 1.3.8. If $X, Y$ are linear spaces, an $f \in \mathcal{L}(X, Y)$ is a linear isomorphism between $X, Y$, if there is $g: Y \rightarrow X$ with $f \circ g=\operatorname{id}_{Y}$ and $g \circ f=\operatorname{id}_{X}$


In this case, we write $f: X \simeq Y$, and we say that the linear spaces $X$ and $Y$ are (linearly) isomorphic.

Next we see that two isomorphic finite-dimensional linear spaces have the same dimension.

Proposition 1.3.9. Let $X, Y$ be linear spaces, and $f \in \mathcal{L}(X, Y)$ a linear isomorphism.
(i) $f$ is a bijection (i.e., an injection and a surjection).
(ii) If $g: Y \rightarrow X$ such that $f \circ g=\operatorname{id}_{Y}$ and $g \circ f=\operatorname{id}_{X}$, then $g \in \mathcal{L}(Y, X)$.
(iii) If $n \in \mathbb{N}$, and $\operatorname{dim}(X)=n$, then $\operatorname{dim}(Y)=n$.
(iv) If $h: X \rightarrow Y$ is a linear map, which is a bijection, then $h$ is a linear isomorphism.

Proof. Exercise.
The condition (iv) above could be taken as the definition of a linear isomorphism. If $n \geq 1$, an $n$-dimensional linear space is isomorphic to $\mathbb{R}^{n}$.

Corollary 1.3.10. If $X$ is a linear space, and $n \geq 1$, then $\operatorname{dim}(X)=n$ if and only if $X$ is isomorphic to $\mathbb{R}^{n}$.

Proof. Exercise.
The set of operators $\mathcal{L}(X)$ of a linear space $X$ is algebraically more interesting than $\mathcal{L}(X, Y)$, since a "multiplication", the composition of functions, is defined between its elements.

Definition 1.3.11. If $X$ is a linear space, and $T \in \mathcal{L}(X)$, we define

$$
T^{n}:= \begin{cases}\operatorname{id}_{X} & , n=0 \\ T \circ T^{n-1} & , n>0\end{cases}
$$

E.g., $T^{3}=T \circ T \circ T$


Remark 1.3.12. If $X$ is a linear space, and $P \in \mathcal{L}(X)$, such that $P^{2}=P$, then

$$
X=\operatorname{Ker}(P) \oplus \operatorname{Im}(P)
$$

Proof. Exercise.
Remark 1.3.13. Let $X$ be a linear space, $T \in \mathcal{L}(X)$, with $T^{2}=\operatorname{id}_{X}$, and let

$$
P:=\frac{1}{2}\left(\mathrm{id}_{X}+T\right) \quad \& \quad Q:=\frac{1}{2}\left(\mathrm{id}_{X}-T\right) .
$$

(i) $P+Q=\operatorname{id}_{X}$.
(ii) $P^{2}=P$, and $Q^{2}=Q$.
(iii) $P Q=Q P=\mathbf{0}$.

Proof. Exercise.

Proposition 1.3.14. Let $n \geq 1, X, Z$ be linear spaces, $Y \subseteq X$, and let the function $f_{0}: Y \rightarrow Z$.
(i) If $X=\langle Y\rangle$, there is at most one linear map $f: X \rightarrow Z$ that extends $f_{0}$ i.e., $f(y)=f_{0}(y)$, for every $y \in Y$, or, in other words, the following diagram commutes

(ii) If $Y=\left\{v_{1}, \ldots, v_{n}\right\}$ is a basis of $X$, there is a unique linear map $f: X \rightarrow Z$ that extends $f_{0}$, and hence, if $g, h: X \rightarrow Z$ are linear maps, we have that ${ }^{3}$

$$
g_{\mid Y}=h_{\mid Y} \Rightarrow g=h .
$$

Proof. (i) If $X$ is a trivial linear space, then $Y=\emptyset$ or $Y=X$. In the first case, $f_{0}$ is the empty set (as a set of pairs), and the only linear map that extends $f_{0}$ is the constant zero linear map. If $Y=X$, the only extension of $f_{0}$ is $f_{0}$ itself. If $X$ is non-trivial, let $f, g: X \rightarrow Z$ be linear maps such that their restrictions $f_{\mid Y}, g_{\mid Y}$ to $Y$ are equal to $f_{0}$, i.e.,

$$
\forall_{y \in Y}\left(f(y)=f_{0}(y)=g(y)\right) .
$$

If $x \in X$, let $a_{1}, \ldots, a_{n} \in \mathbb{R}$ and $y_{1}, \ldots y_{n} \in Y$ such that $x=\sum_{i=1}^{n} a_{i} y_{i}$. By the Remark 1.3.6(v) we have that

$$
f(x)=f\left(\sum_{i=1}^{n} a_{i} y_{i}\right)=\sum_{i=1}^{n} a_{i} f\left(y_{i}\right)=\sum_{i=1}^{n} a_{i} g\left(y_{i}\right)=g\left(\sum_{i=1}^{n} a_{i} y_{i}\right)=g(x)
$$

(ii) If $x \in X$, then $x$ has a unique writing as $x=\sum_{=1}^{n} a_{i} v_{i}$, for some $a_{1}, \ldots, a_{n} \in \mathbb{R}$. We define $f: X \rightarrow Z$ by

$$
f\left(\sum_{i=1}^{n} a_{i} v_{i}\right)=\sum_{i=1}^{n} a_{i} f_{0}\left(v_{i}\right) .
$$

It is easy to check that $f$ is a linear map that extends $f_{0}$. Since $Y$ generates $X$, by the case (i) we get that $f$ is the unique extension of $f_{0}$. Moreover, if $g$ and $h$ are equal on the basis $Y$, then they are equal as functions from $X$ to $Z$, since there is a unique extension of the restriction $g_{\mid Y}$ of $g$ to $Y$.

[^2]
### 1.4. The space of matrices

The set of $m \times n$-matrices $M_{m, n}(\mathbb{R})$, and the set of square matrices $M_{n}(\mathbb{R}):=$ $M_{n, n}(\mathbb{R})$ was defined in the Definition 1.3.4.

Remark 1.4.1. $M_{m, n}(\mathbb{R})$ is a linear space of dimension $m n$.
Proof. The fact that $M_{m, n}(\mathbb{R})$ is a linear space is immediate from the Definition 1.3.4. To determine the dimension of $M_{m, n}(\mathbb{R})$, we associate to an $m \times n$-matrix

$$
A:=\left[\begin{array}{ccc}
a_{11} & \ldots & a_{1 n} \\
\vdots & \vdots & \vdots \\
a_{i 1} & \ldots & a_{i n} \\
\vdots & \vdots & \vdots \\
a_{m 1} & \ldots & a_{m n}
\end{array}\right]
$$

the following element of $\mathbb{R}^{m n}$

$$
\left(a_{11}, \ldots, a_{1 n}, \ldots, a_{i 1}, \ldots, a_{i n}, \ldots, a_{m 1}, \ldots, a_{m n}\right)
$$

E.g., to the $2 \times 2$-matrix

$$
\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right]
$$

we associate the 4-tuple

$$
(a, b, c, d) .
$$

It is easy to see that this mapping $e: M_{m, n}(\mathbb{R}) \rightarrow \mathbb{R}^{m n}$ is a linear isomorphism, hence by the Proposition 1.3.9(iii) we get $\operatorname{dim}\left(M_{m, n}(\mathbb{R})\right)=\operatorname{dim}\left(\mathbb{R}^{m n}\right)=m n$.

Definition 1.4.2. Let the mapping ${ }^{t}: M_{m, n}(\mathbb{R}) \rightarrow M_{n, m}(\mathbb{R})$, defined by

$$
\left[a_{i j}\right] \mapsto\left[a_{i j}\right]^{t},
$$

where

$$
\left[a_{i j}\right]^{t}:=\left[b_{j i}\right], \quad b_{j i}:=a_{i j} .
$$

The matrix $\left[a_{i j}\right]^{t}$ is called the transpose of $\left[a_{i j}\right]$, and it has columns the rows of $\left[a_{i j}\right]$ and rows the columns of $\left[a_{i j}\right]$. If $A \in M_{n}(\mathbb{R})$ with $A^{t}=A$, we say that $A$ is symmetric, and we denote their set by $\operatorname{Sym}_{n}(\mathbb{R})$. A diagonal matrix in $M_{n}(\mathbb{R})$ has the form

$$
\left[\begin{array}{cccc}
\lambda_{1} & & & \\
& \lambda_{2} & & \\
& & \ddots & \\
& & & \lambda_{n}
\end{array}\right]:=\left[\begin{array}{cccc}
\lambda_{1} & 0 & \ldots & 0 \\
0 & \lambda_{2} & \ldots & 0 \\
\vdots & \vdots & \ddots & \vdots \\
0 & 0 & \ldots & \lambda_{n}
\end{array}\right]=: \operatorname{Diag}\left(\lambda_{1}, \ldots, \lambda_{n}\right)
$$

We denote by $I_{n}$ the unit matrix in $M_{n}(\mathbb{R})$, defined by

$$
I_{n}:=\left[\begin{array}{llll}
1 & & & \\
& 1 & & \\
& & \ddots & \\
& & & 1
\end{array}\right]=:\left[\delta_{i j}\right]
$$

where ${ }^{4}$

$$
\delta_{i j}:= \begin{cases}1 & , \text { if } i=j \\ 0 & , \text { if } i \neq j\end{cases}
$$

E.g., if we consider the $2 \times 3$-matrix

$$
A:=\left[\begin{array}{lll}
2 & 1 & 0 \\
1 & 3 & 5
\end{array}\right]
$$

then its transpose $A^{t}$ is the following $3 \times 2$-matrix

$$
A^{t}:=\left[\begin{array}{ll}
2 & 1 \\
1 & 3 \\
0 & 5
\end{array}\right]
$$

An example of a symmetric matrix is the following:

$$
A=\left[\begin{array}{rrr}
3 & 1 & -2 \\
1 & 5 & 4 \\
-2 & 4 & -8
\end{array}\right]=A^{t}
$$

Remark 1.4.3. Let $A, B \in M_{m, n}(\mathbb{R}), C \in M_{n}(\mathbb{R})$, and $a \in \mathbb{R}$.
(i) $(A+B)^{t}=A^{t}+B^{t}$.
(ii) $(a \cdot B)^{t}=a \cdot B^{t}$.
(iii) $\left(A^{t}\right)^{t}=A$.
(iii) $C+C^{t}$ is symmetric.

Proof. Exercise.
Next we define the multiplication between matrices, an operation which, as we shall see later, is related to the composition of linear maps. To define the multiplication $A B$ the number of columns of $A$ has to be the number of rows of $B$ !

Definition 1.4.4. If $A:=\left[a_{i j}\right] \in M_{m, n}(\mathbb{R})$ and $B:=\left[b_{j k}\right] \in M_{n, l}(\mathbb{R})$, their product $A B \in M_{m, l}(\mathbb{R})$ is defined by

$$
\begin{gathered}
A B:=\left[a_{i j}\right]\left[b_{j k}\right]:=\left[c_{i k}\right], \\
c_{i k}:=\sum_{j=1}^{n} a_{i j} b_{j k}
\end{gathered}
$$

[^3]for every $1 \leq i \leq m$ and $1 \leq k \leq l$. If $A \in M_{n}(\mathbb{R})$, let
\[

A^{n}:= $$
\begin{cases}I_{n} & , n=0 \\ A A^{n-1} & , n>0\end{cases}
$$
\]

A matrix $A \in M_{n}(\mathbb{R})$ is invertible, if there is $B \in M_{n}(\mathbb{R})$ such that $A B=B A=I_{n}$. We denote by $\operatorname{Inv}_{n}(\mathbb{R})$ the set of invertible matrices in $M_{n}(\mathbb{R})$.
E.g., if

$$
A:=\left[\begin{array}{lll}
2 & 1 & 5 \\
1 & 3 & 2
\end{array}\right] \quad \& \quad B:=\left[\begin{array}{rr}
3 & 4 \\
-1 & 2 \\
2 & 1
\end{array}\right]
$$

then

$$
A B:=\left[\begin{array}{lll}
2 & 1 & 5 \\
1 & 3 & 2
\end{array}\right]\left[\begin{array}{rr}
3 & 4 \\
-1 & 2 \\
2 & 1
\end{array}\right]=\left[\begin{array}{cc}
15 & 15 \\
4 & 12
\end{array}\right]
$$

It is not always true that $A B=B A$. E.g.,

$$
\left[\begin{array}{ll}
3 & 2 \\
0 & 1
\end{array}\right]\left[\begin{array}{rr}
2 & -1 \\
0 & 5
\end{array}\right]=\left[\begin{array}{ll}
6 & 7 \\
0 & 5
\end{array}\right]
$$

and

$$
\left[\begin{array}{rr}
2 & -1 \\
0 & 5
\end{array}\right]\left[\begin{array}{ll}
3 & 2 \\
0 & 1
\end{array}\right]=\left[\begin{array}{rr}
6 & -3 \\
0 & 5
\end{array}\right]
$$

If $a, b \in \mathbb{R}$, and

$$
A:=\left[\begin{array}{ll}
1 & a \\
0 & 1
\end{array}\right] \quad \& \quad B:=\left[\begin{array}{ll}
1 & b \\
0 & 1
\end{array}\right]
$$

then

$$
A B:=\left[\begin{array}{cc}
1 & a+b \\
0 & 1
\end{array}\right]
$$

Hence

$$
\left[\begin{array}{rr}
1 & -a \\
0 & 1
\end{array}\right]\left[\begin{array}{ll}
1 & a \\
0 & 1
\end{array}\right]=I_{2}
$$

Notice that, in contrast to what happens in $\mathbb{R}$, there are non-zero square matrices that are not invertible, like the matrix

$$
\left[\begin{array}{ll}
1 & 1 \\
1 & 1
\end{array}\right]
$$

Proposition 1.4.5. Let $A \in M_{m, n}(\mathbb{R}), B, C \in M_{n, l}(\mathbb{R})$, and $D \in M_{l, s}(\mathbb{R})$.
(i) $A I_{n}=A$ and $I_{m} A=A$.
(ii) $A(B+C)=A B+A C$.
(iii) If $a \in \mathbb{R}$, then $A(a \cdot B)=a \cdot(A B)$.
(iv) $A(B D)=(A B) D$.
(v) The multiplication $B^{t} A^{t}$ is well-defined, and $(A B)^{t}=B^{t} A^{t}$.

Proof. Exercise.
Corollary 1.4.6. Let $A, B, C \in M_{n}(\mathbb{R})$.
(i) If $A B=B A=I_{n}=A C=C A$, then $B=C$. We denote the unique matrix $B$ such that $A B=B A=I_{n}$ by $A^{-1}$, and we call it the inverse of $A$.
(ii) $I_{n}^{t}=I_{n}$.
(ii) If $A$ is invertible, then $\left(A^{-1}\right)^{t}=\left(A^{t}\right)^{-1}$.

Proof. (i) $C=I_{n} C=(A B) C=(B A) C=B(A C)=B I_{n}=B$.
(ii) $\left[\delta_{i j}\right]^{t}:=\left[d_{i j}\right]$, where $d_{i j}:=\delta_{i j}$, and what we want follows from the obvious equality $\delta_{i j}=\delta_{j i}$.
(iii) By the Proposition 1.4.5(v) and the case (ii) we have that $I_{n}=I_{n}^{t}=\left(A A^{-1}\right)^{t}=$ $\left(A^{-1}\right)^{t} A^{t}$, and $I_{n}=I_{n}^{t}=\left(A^{-1} A\right)^{t}=A^{t}\left(A^{-1}\right)^{t}$. Since $I_{n}=\left(A^{t}\right)^{-1} A^{t}=A^{t}\left(A^{t}\right)^{-1}$, by the case (i) we get $\left(A^{-1}\right)^{t}=\left(A^{t}\right)^{-1}$.

One can show that if $A, B \in M_{n}(\mathbb{R})$, then

$$
A B=I_{n} \Rightarrow B A=I_{n}
$$

hence we do not need to check both equalities in order to show that a matrix $A$ is invertible. Note that this is the case only when the product $A B$ is equal to $I_{n}$. If $A, B \in M_{n}(\mathbb{R})$ are invertible, then $A B$ is also invertible and $(A B)^{-1}=B^{-1} A^{-1}$, since

$$
(A B)\left(B^{-1} A^{-1}\right)=A\left[B\left(B^{-1} A^{-1}\right)\right]=A\left[\left(B B^{-1}\right) A^{-1}\right]=A\left[I_{n} A^{-1}\right]=A A^{-1}=I_{n}
$$

### 1.5. Matrices and linear maps

Matrices can be used to represent linear maps. Let's see the following important example. If $\theta \in \mathbb{R}$, let the matrix

$$
R(\theta):=\left[\begin{array}{rr}
\cos \theta & -\sin \theta \\
\sin \theta & \cos \theta
\end{array}\right] .
$$

Let the $\operatorname{map} R_{\theta}: \mathbb{R}^{2} \rightarrow \mathbb{R}^{2}$ defined by

$$
\begin{aligned}
R_{\theta}(x, y) & :=\left[\begin{array}{rr}
\cos \theta & -\sin \theta \\
\sin \theta & \cos \theta
\end{array}\right]\left[\begin{array}{l}
x \\
y
\end{array}\right] \\
& =\left[\begin{array}{rr}
\cos \theta & -\sin \theta \\
\sin \theta & \cos \theta
\end{array}\right]\left[\begin{array}{r}
r \cos \phi \\
r \sin \phi
\end{array}\right] \\
& =r\left[\begin{array}{l}
\cos \theta \cos \phi-\sin \theta \sin \phi \\
\sin \theta \cos \phi+\cos \theta \cos \phi
\end{array}\right] \\
& =r\left[\begin{array}{l}
\cos (\theta+\phi) \\
\sin (\theta+\phi)
\end{array}\right]
\end{aligned}
$$


where $r:=\sqrt{x^{2}+y^{2}}$. Hence, $R_{\theta}$ is the anti-clockwise $\theta$-rotation of the vector $(x, y)$. If $\theta_{1}, \theta_{2} \in \mathbb{R}$, it is easy to see that

$$
R\left(\theta_{1}\right) R\left(\theta_{2}\right)=R\left(\theta_{1}+\theta_{2}\right)
$$

From that we can infer that the matrix $R(\theta)$ has an inverse.
Definition 1.5.1. If $A:=\left[a_{i j}\right] \in M_{m, n}(\mathbb{R})$, the linear map of $A$ is the mapping

$$
\begin{aligned}
& T_{A}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m} \\
& T_{A}(X):=A X
\end{aligned}
$$

where we view an arbitrary element $x:=\left(x_{1}, \ldots, x_{n}\right) \in \mathbb{R}^{n}$ as an $n \times 1$-matrix $X$ and the output $m \times 1$-matrix represents a vector in $\mathbb{R}^{m}$. I.e., we have

$$
\left[\begin{array}{c}
T_{A}(X)_{1} \\
\vdots \\
T_{A}(X)_{m}
\end{array}\right]:=\left[\begin{array}{ccc}
a_{11} & \ldots & a_{1 n} \\
\vdots & \vdots & \vdots \\
a_{m 1} & \ldots & a_{m n}
\end{array}\right]\left[\begin{array}{c}
x_{1} \\
\vdots \\
x_{n}
\end{array}\right] .
$$

In a non-matrix form we write

$$
T_{A}(x):=\left(\sum_{j=1}^{n} a_{1 j} x_{j}, \ldots, \sum_{j=1}^{n} a_{i j} x_{j}, \ldots, \sum_{j=1}^{n} a_{m j} x_{j}\right) .
$$

If $\left\{e_{1}, \ldots, e_{n}\right\}$ is the standard basis of $\mathbb{R}^{n}$, and $l \in\{1, \ldots, n\}$, then

$$
T_{A}\left(e_{l}\right):=\left(a_{1 l}, \ldots, a_{m l}\right)=A^{l}
$$

where $A^{l}$ is the $l$-column of the matrix $A$. and hence

$$
T_{A}\left(e_{l}\right)_{i}=a_{i l}
$$

for every $i \in\{1, \ldots, m\}$. Using the Proposition 1.4.5 we can show the following.

Proposition 1.5.2. If $A, B \in M_{m, n}(\mathbb{R})$, and $a \in \mathbb{R}$, the following hold:
(i) $T_{A} \in \mathcal{L}\left(\mathbb{R}^{n}, \mathbb{R}^{m}\right)$.
(ii) If $T_{A}=\mathbf{0}$, then $A=\mathbf{0}_{m n}$, and if $T_{A}=T_{B}$, then $A=B$.
(iii) $T_{A+B}=T_{A}+T_{B}$.
(iv) $T_{a \cdot A}=a T_{A}$.
(v) $T_{I_{n}}=\mathrm{id}_{\mathbb{R}_{n}}$ and $T_{\mathbf{0}_{m n}}=\mathbf{0}$.
(vi) If $C \in M_{n, l}(\mathbb{R})$, then $T_{A C}=T_{A} \circ T_{C}$

(vii) If $A$ is invertible, then $T_{A}$ is invertible and $T_{A}^{-1}=T_{A^{-1}}$.
(viii) The function $\mathcal{T}: M_{m, n}(\mathbb{R}) \rightarrow \mathcal{L}\left(\mathbb{R}^{n}, \mathbb{R}^{m}\right)$, defined by $A \mapsto T_{A}$, is a linear map.

Proof. Exercise.
So far we defined a a linear map $T_{A}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$, given a matrix $A \in M_{m, n}(\mathbb{R})$. Next we define a matrix $A_{T} \in M_{m, n}(\mathbb{R})$, given a linear map $T: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$. The two constructions are inverse to each other.

THEOREM 1.5.3. Let $n, m \geq 1$. If $T: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ is a linear map, there is a matrix $A_{T} \in M_{m, n}(\mathbb{R})$ such that $T=T_{A_{T}}$ i.e., for every $x \in \mathbb{R}^{n}$ we have that

$$
T(x)=T_{A_{T}}(x):=A_{T} x
$$

The matrix $A_{T}$ is called the matrix of the linear map $T$.
Proof. If $B:=\left\{e_{1}, \ldots, e_{n}\right\}$ is the standard basis of $\mathbb{R}^{n}$, then for every $i \in$ $\{1, \ldots, n\}$ we write $T\left(e_{1}\right)$ a linear combination of the standard basis of $\mathbb{R}^{m}$ i.e.,

$$
T\left(e_{i}\right):=\left(T\left(e_{i}\right)_{1}, \ldots, T\left(e_{i}\right)_{m}\right)
$$

The matrix $A_{T}$ is formed by taking these $m$-tuples as its columns i.e., we define

$$
A_{T}:=\left[\begin{array}{ccc}
T\left(e_{1}\right)_{1} & \ldots & T\left(e_{n}\right)_{1} \\
\vdots & \vdots & \vdots \\
T\left(e_{1}\right)_{j} & \ldots & T\left(e_{n}\right)_{j} \\
\vdots & \vdots & \vdots \\
T\left(e_{1}\right)_{m} & \ldots & T\left(e_{n}\right)_{m}
\end{array}\right]=:\left[a_{j i}\right]=\left[T\left(e_{i}\right)_{j}\right]
$$

By the Proposition 1.3.14, to show that the linear maps $T$ and $T_{A_{T}}$ are equal, it suffices to show that they are equal on the elements of $B$. Since

$$
T_{A_{T}}\left(e_{i}\right):=A_{T} e_{i}:=\left[T\left(e_{i}\right)_{j}\right] e_{i}=\left[a_{j i}\right] e_{i}=\left[c_{j 1}\right],
$$

where

$$
c_{j 1}=\sum_{i=1}^{n} a_{j i} b_{i 1}=a_{j i}:=T\left(e_{i}\right)_{j},
$$

we get ${ }^{5}$ the required equality with the vector $T\left(e_{i}\right):=\left(T\left(e_{i}\right)_{1}, \ldots, T\left(e_{i}\right)_{m}\right)$.
For example, if $T: \mathbb{R}^{2} \rightarrow \mathbb{R}^{2}$ is a linear map such that

$$
T(0,1):=(a, c) \quad \& \quad T(1,0):=(b, d)
$$

then we have that

$$
A_{T}=\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right]
$$

Proposition 1.5.4. Let the function $\mathcal{A}: L\left(\mathbb{R}^{n}\right) \rightarrow M_{n}(\mathbb{R})$

$$
T \mapsto A_{T}:=\mathcal{A}(T) .
$$

(i) The mappings $\mathcal{T}$ and $\mathcal{A}$ satisfy the following conditions:
(i) $\mathcal{A} \circ \mathcal{T}=\mathrm{id}_{M_{n}(\mathbb{R})}$ and $\mathcal{T} \circ \mathcal{A}=\mathrm{id}_{L\left(\mathbb{R}^{n}\right)}$

(ii) $A_{S \circ T}=A_{S} A_{T}$.
(iii) $A_{I_{n}}=I_{n}$ and $A_{O_{n}}=O_{n}$.
(ix) $A_{S+T}=A_{S}+A_{T}$.
(x) $A_{\lambda T}=\lambda A_{T}$.
(xi) If $T$ is invertible, then $A_{T}$ is invertible and $A_{T}^{-1}=A_{T^{-1}}$.

Proof. Exercise.

### 1.6. Determinants

Definition 1.6.1. If

$$
A=\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right]
$$

[^4]is a $2 \times 2$-matrix, its determinant $\operatorname{Det}(A)$ is defined by
\[

\operatorname{Det}(A):=\left|$$
\begin{array}{ll}
a & b \\
c & d
\end{array}
$$\right|:=a d-b c
\]

If

$$
A^{1}:=\left[\begin{array}{l}
a \\
c
\end{array}\right] \quad \& \quad A^{2}:=\left[\begin{array}{l}
b \\
d
\end{array}\right]
$$

are the columns of $A$, we use the notation

$$
\operatorname{Det}(A)=\operatorname{Det}\left(A^{1}, A^{2}\right)
$$

We have that

$$
\operatorname{Det}\left(I_{2}\right):=\left|\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right|:=1-0=1
$$

It is also clear that

$$
\operatorname{Det}(A):=\left|\begin{array}{ll}
a & b \\
c & d
\end{array}\right|:=a d-b c=:\left|\begin{array}{ll}
a & c \\
b & d
\end{array}\right|=: \operatorname{Det}\left(A^{t}\right)
$$

REmARK 1.6.2. Let the following $2 \times 1$ matrices:

$$
A^{1}:=\left[\begin{array}{l}
a_{1} \\
a_{2}
\end{array}\right], \quad C^{1}:=\left[\begin{array}{l}
c_{1} \\
c_{2}
\end{array}\right], \quad B^{2}:=\left[\begin{array}{l}
b_{1} \\
b_{2}
\end{array}\right], \quad D^{2}:=\left[\begin{array}{l}
d_{1} \\
d_{2}
\end{array}\right]
$$

The following hold.
(i) $\operatorname{Det}\left(A^{1}+C^{1}, B^{2}\right)=\operatorname{Det}\left(A^{1}, B^{2}\right)+\operatorname{Det}\left(C^{1}, B^{2}\right)$.
(ii) $\operatorname{Det}\left(A^{1}, B^{2}+D^{2}\right)=\operatorname{Det}\left(A^{1}, B^{2}\right)+\operatorname{Det}\left(A^{1}, D^{2}\right)$.
(iii) If $\lambda \in \mathbb{R}$, then $\operatorname{Det}\left(\lambda A^{1}, B^{2}\right)=\lambda \operatorname{Det}\left(A^{1}, B^{2}\right)=\operatorname{Det}\left(A^{1}, \lambda B^{2}\right)$.
(iv) If $A^{1}=B^{2}$, then $\operatorname{Det}\left(A^{1}, B^{2}\right)=0$.

Proof. We prove only (i), and the rest is an exercise.

$$
\begin{aligned}
\operatorname{Det}\left(A^{1}+C^{1}, B^{2}\right) & :=\left|\begin{array}{ll}
a_{1}+c_{1} & b_{1} \\
a_{2}+c_{2} & b_{2}
\end{array}\right| \\
& :=\left(a_{1}+c_{1}\right) b_{2}-b_{1}\left(a_{2}+c_{2}\right) \\
& =\left(a_{1} b_{2}-b_{1} a_{2}\right)+\left(c_{1} b_{2}-b_{1} c_{2}\right) \\
& :=\left|\begin{array}{ll}
a_{1} & b_{1} \\
a_{2} & b_{2}
\end{array}\right|+\left|\begin{array}{ll}
c_{1} & b_{1} \\
c_{2} & b_{2}
\end{array}\right| \\
& :=\operatorname{Det}\left(A^{1}, B^{2}\right)+\operatorname{Det}\left(C^{1}, B^{2}\right)
\end{aligned}
$$

Although one can use the definition of $\operatorname{Det}(A)$ to show the following corollary, its proof is simpler, if we use the fundamental properties of the Remark 1.6.2.

Corollary 1.6.3. Let the following $2 \times 1$ matrices:

$$
A^{1}:=\left[\begin{array}{l}
a_{1} \\
a_{2}
\end{array}\right], \quad B^{2}:=\left[\begin{array}{l}
b_{1} \\
b_{2}
\end{array}\right] .
$$

The following hold.
(i) If $\lambda \in \mathbb{R}$, then $\operatorname{Det}\left(A^{1}+\lambda B^{2}, B^{2}\right)=\operatorname{Det}\left(A^{1}, B^{2}\right)$.
(ii) If $\lambda \in \mathbb{R}$, then $\operatorname{Det}\left(A^{1}, B^{2}+\lambda A^{1}\right)=\operatorname{Det}\left(A^{1}, B^{2}\right)$.
(iii) $\operatorname{Det}\left(A^{1}, B^{2}\right)=-\operatorname{Det}\left(B^{2}, A^{1}\right)$.

Proof. Exercise.
The determinant of a matrix $A$ provides non-trivial information on vectors related to $A$. We have seen that $\operatorname{Det}\left(I_{2}\right)=1 \neq 0$, and we know that the columns $e_{1}:=(1,0)$ and $e_{2}:=(0,1)$ of the matrix $I_{2}$ are linearly independent elements. This is a special case of the following general fact.

Proposition 1.6.4. Let the following $2 \times 1$ matrices:

$$
A:=\left[\begin{array}{l}
a_{1} \\
a_{2}
\end{array}\right], \quad B:=\left[\begin{array}{l}
b_{1} \\
b_{2}
\end{array}\right] .
$$

The vectors $\left(a_{1}, a_{2}\right)$ and $\left(b_{1}, b_{2}\right)$ are linearly independent in $\mathbb{R}^{2}$ if and only if

$$
\operatorname{Det}(A, B) \neq 0
$$

Proof. $(\Rightarrow)$ Suppose that $\left(a_{1}, a_{2}\right)$ and $\left(b_{1}, b_{2}\right)$ are linearly independent in $\mathbb{R}^{2}$, and suppose that

$$
\operatorname{Det}(A, B):=\left|\begin{array}{ll}
a_{1} & b_{1} \\
a_{2} & b_{2}
\end{array}\right|:=a_{1} b_{2}-b_{1} a_{2}=0
$$

Since then we have that

$$
b_{2}\left(a_{1}, a_{2}\right)+\left(-a_{2}\right)\left(b_{1}, b_{2}\right)=\left(b_{2} a_{1}-a_{2} b_{1}, b_{2} a_{2}-a_{2} b_{2}\right)=(0,0)
$$

by the hypothesis of linear independence of $\left(a_{1}, a_{2}\right)$ and $\left(b_{1}, b_{2}\right)$ we get

$$
b_{2}=0=-a_{2}=a_{2} .
$$

Hence the two vectors take the form $\left(a_{1}, 0\right)$ and $\left(b_{1}, 0\right)$. Since they are linearly independent, these are non-zero vectors, hence $a_{1} \neq 0$ and $b_{1} \neq 0$. Consequently, we have that $\left(a_{1}, 0\right)=\frac{a_{1}}{b_{1}}\left(b_{1}, 0\right)$ i.e., the vectors $\left(a_{1}, a_{2}\right)$ and $\left(b_{1}, b_{2}\right)$ are linearly dependent, which is a contradiction. Hence, $\operatorname{Det}(A, B) \neq 0$.
$(\Leftarrow)$ Suppose that $\operatorname{Det}(A, B) \neq 0$, and let $\lambda, \mu \in \mathbb{R}$ such that

$$
\lambda\left(a_{1}, a_{2}\right)+\mu\left(b_{1}, b_{2}\right)=(0,0) \Leftrightarrow\left(\lambda a_{1}+\mu b_{1}, \lambda a_{2}+\mu b_{2}\right)=(0,0)
$$

hence

$$
\lambda a_{1}=-\mu b_{1} \quad \& \quad \lambda a_{2}=-\mu b_{2}
$$

Suppose that $\lambda \neq 0$ (if we suppose that $\mu \neq 0$. we proceed similarly). By the Remark 1.6.2 we have that

$$
\begin{aligned}
\operatorname{Det}(A, B) & =\left|\begin{array}{cc}
\left(\frac{-\mu}{\lambda}\right) b_{1} & b_{1} \\
\left(\frac{-\mu}{\lambda}\right) b_{2} & b_{2}
\end{array}\right| \\
& =\left(\frac{-\mu}{\lambda}\right)\left|\begin{array}{ll}
b_{1} & b_{1} \\
b_{2} & b_{2}
\end{array}\right| \\
& =\left(\frac{-\mu}{\lambda}\right) 0 \\
& =0
\end{aligned}
$$

which is a contradiction. Hence $\lambda=0=\mu$, and the vectors $\left(a_{1}, a_{2}\right),\left(b_{1}, b_{2}\right)$ are linearly independent.

Proposition 1.6.5. Let $A, B \in M_{2}(\mathbb{R})$.
(i) $\operatorname{Det}(A B)=\operatorname{Det}(A) \operatorname{Det}(B)$.
(ii) $A$ is invertible if and only if $\operatorname{Det}(A) \neq 0$.

Proof. (i) Exercise.
(ii) If $A A^{-1}=I_{2}$, then by the case (i) we have that

$$
1=\operatorname{Det}\left(I_{2}\right)=\operatorname{Det}\left(A A^{-1}\right)=\operatorname{Det}(A) \operatorname{Det}\left(A^{-1}\right),
$$

hence $\operatorname{Det}(A) \neq 0, \operatorname{Det}\left(A^{-1}\right) \neq 0$, and

$$
\operatorname{Det}\left(A^{-1}\right)=\frac{1}{\operatorname{Det}(A)}
$$

For the converse let

$$
A=\left[\begin{array}{ll}
a & b \\
c & d
\end{array}\right]
$$

and suppose that

$$
\operatorname{Det}(A):=\left|\begin{array}{ll}
a & b \\
c & d
\end{array}\right|:=a d-b c \neq 0
$$

We show that the system

$$
\begin{gathered}
{\left[\begin{array}{cc}
a & b \\
c & d
\end{array}\right]\left[\begin{array}{cc}
x & y \\
z & w
\end{array}\right]=\left[\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right] \Leftrightarrow} \\
{\left[\begin{array}{cc}
a x+b z & a y+b w \\
c x+d z & c y+d w
\end{array}\right]=\left[\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right] \Leftrightarrow} \\
a x+b z=1 \quad \& \quad c x+d z=0,
\end{gathered}
$$

and

$$
a y+b w=0 \quad \& \quad c y+d w=1
$$

has a solution. If we multiply the equation $a x+b z=1$ by $d$ and the equation $c x+d z=0$ by $b$, and we subtract them we get

$$
d a x+d b z-b c x-b d z=d \Leftrightarrow x=\frac{d}{a d-b c}
$$

Working similarly, we get

$$
A^{-1}:=\left[\begin{array}{cc}
x & y \\
z & w
\end{array}\right]=\frac{1}{\operatorname{Det}(A)}\left[\begin{array}{rr}
d & -b \\
-c & a
\end{array}\right]
$$

Definition 1.6.6. If

$$
A=\left[\begin{array}{lll}
a_{11} & a_{12} & a_{13} \\
a_{21} & a_{22} & a_{23} \\
a_{31} & a_{32} & a_{33}
\end{array}\right]
$$

is a $3 \times 3$-matrix, its determinant $\operatorname{Det}(A)$ is defined by

$$
\operatorname{Det}(A):=\left|\begin{array}{lll}
a_{11} & a_{12} & a_{13} \\
a_{21} & a_{22} & a_{23} \\
a_{31} & a_{32} & a_{33}
\end{array}\right|:=a_{11}\left|\begin{array}{ll}
a_{22} & a_{23} \\
a_{32} & a_{33}
\end{array}\right|-a_{12}\left|\begin{array}{cc}
a_{21} & a_{23} \\
a_{31} & a_{33}
\end{array}\right|+a_{13}\left|\begin{array}{ll}
a_{21} & a_{22} \\
a_{31} & a_{32}
\end{array}\right|
$$

As expected, we have that

$$
\operatorname{Det}\left(I_{3}\right):=\left|\begin{array}{lll}
1 & 0 & 0 \\
0 & 1 & 0 \\
0 & 0 & 1
\end{array}\right|:=1\left|\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right|-0\left|\begin{array}{ll}
0 & 0 \\
0 & 1
\end{array}\right|+0\left|\begin{array}{ll}
0 & 1 \\
0 & 0
\end{array}\right|=1
$$

More generally, if we consider a matrix in diagonal form, then for the corresponding determinant we have that

$$
\left|\begin{array}{ccc}
\lambda_{1} & 0 & 0 \\
0 & \lambda_{2} & 0 \\
0 & 0 & \lambda_{3}
\end{array}\right|:=\lambda_{1}\left|\begin{array}{cc}
\lambda_{2} & 0 \\
0 & \lambda_{3}
\end{array}\right|-0\left|\begin{array}{cc}
0 & 0 \\
0 & \lambda_{3}
\end{array}\right|+0\left|\begin{array}{cc}
0 & \lambda_{2} \\
0 & 0
\end{array}\right|=\lambda_{1} \lambda_{2} \lambda_{3} .
$$

All results we showed for the determinant of a matrix in $M_{2}(\mathbb{R})$ hold also for the determinant of a matrix in $M_{3}(\mathbb{R})$.

### 1.7. The inner product on $\mathbb{R}^{n}$

Definition 1.7.1. Let $X$ be a linear space. An inner product on $X$ is a mapping $\langle\cdot, \cdot\rangle: X \times X \rightarrow \mathbb{R}$ such that for every $x, y, z \in X$ and $\lambda \in \mathbb{R}$ the following conditions hold:
(i) $\langle x, x\rangle \geq 0$ (positivity).
(ii) $\langle x, x\rangle=0 \Rightarrow x=\mathbf{0}$ (definiteness).
(iii) $\langle x, y\rangle=\langle y, x\rangle$ (symmetry).
(iv) $\langle x+y, z\rangle=\langle x, z\rangle+\langle y, z\rangle$ (left additivity).
(v) $\langle\lambda x, y\rangle=\lambda\langle x, y\rangle$ (left homogeneous).

If $\langle\cdot, \cdot\rangle$ is an inner product on $X$, the pair $(X,\langle\cdot, \cdot\rangle)$ is called an inner product space. A norm on $X$ is a mapping $\|\|:. X \rightarrow \mathbb{R}$ such that for every $x, y \in X$ and $\lambda \in \mathbb{R}$ the following hold:
(i) $\|x\| \geq 0$ (positivity).
(ii) $\|x\|=0 \Rightarrow x=\mathbf{0}$ (definiteness).
(iii) $\|x+y\| \leq\|x\|+\|y\|$ (triangle inequality).
(iv) $\|\lambda x\|=|\lambda|\|x\|$.

If $\|$.$\| is a norm on X$, the pair $(X,\|\cdot\|)$ is called a normed space.
Because of symmetry an inner product is bilinear i.e., it is also right additive and right homogeneous:
(iv') $\langle x, y+z\rangle=\langle x, y\rangle+\langle x, z\rangle$ (right additivity).
$\left(v^{\prime}\right)\langle x, \lambda y\rangle=\lambda\langle x, y\rangle$ (right homogeneous).
Notice also that

$$
\|-x\|=\|(-1) x\|=|-1\|\mid x\|=1\|x\|=\|x\| .
$$

Definition 1.7.2. If $x=\left(x_{1}, \ldots, x_{n}\right)$ and $y=\left(y_{1}, \ldots, y_{n}\right)$ are in $\mathbb{R}^{n}$, their Euclidean inner product is defined by

$$
\langle x, y\rangle:=\sum_{i=1}^{n} x_{i} y_{i}
$$

If $n=1$, then the Euclidean inner product on $\mathbb{R}$ is the standard product on $\mathbb{R}$. By definition we have that

$$
\langle x, x\rangle:=\sum_{i=1}^{n} x_{i} x_{i}=\sum_{i=1}^{n} x_{i}^{2}=x_{1}^{2}+\ldots+x_{n}^{2}
$$

It is easy to see that the Euclidean inner product is an inner product on $\mathbb{R}^{n}$. Next we show that an inner product is determined by its diagonal entries.

Proposition 1.7.3. Let $(X,\langle\cdot, \cdot\rangle)$ be an inner product space and $x, y \in X$.
(i) (Polarization identity) $\langle x, y\rangle=\frac{1}{4}(\langle x+y, x+y\rangle-\langle x-y, x-y\rangle)$.
(ii) $x=\mathbf{0} \Leftrightarrow \forall_{z \in X}(\langle x, z\rangle=0)$.
(iii) $\forall_{z \in X}(\langle x, z\rangle=\langle y, z\rangle) \Rightarrow x=y$.

Proof. Exercise.
If $x=\mathbf{0}$, then $\|x\|=0$, since

$$
\|\mathbf{0}\|=\|0 \cdot \mathbf{0}\|=|0|\|\mathbf{0}\|=0\|\mathbf{0}\|=0
$$

Moreover, if $x=\mathbf{0}$, or $y=\mathbf{0}$, or $y=\lambda x$, for some $\lambda>0$, then the equality holds in the triangle inequality $\|x+y\| \leq\|x\|+\|y\|$.

Definition 1.7.4. If $x \in \mathbb{R}^{n}$, the Euclidean norm $|x|$ of $x$ is defined by

$$
|x|:=\left(\sum_{i=1}^{n} x_{i}^{2}\right)^{\frac{1}{2}}=\sqrt{x_{1}^{2}+x_{2}^{2}+\ldots+x_{n}^{2}}=\sqrt{\langle x, x\rangle} .
$$

Geometrically, if $x \in \mathbb{R}^{n}$, then $|x|$ is the length of the vector $x$. To show that the Euclidean norm is a norm we need the following inequality.

Proposition 1.7.5 (Inequality of Cauchy). If $x, y \in \mathbb{R}^{n}$, then

$$
|\langle x, y\rangle| \leq|x||y| .
$$

Proof. By definition we need to show

$$
\left|\sum_{i=1}^{n} x_{i} y_{i}\right| \leq\left(\sum_{i=1}^{n} x_{i}^{2}\right)^{\frac{1}{2}}\left(\sum_{i=1}^{n} y_{i}^{2}\right)^{\frac{1}{2}}
$$

which is equivalent to

$$
A:=\left(\sum_{i=1}^{n} x_{i} y_{i}\right)^{2} \leq\left(\sum_{i=1}^{n} x_{i}^{2}\right)\left(\sum_{i=1}^{n} y_{i}^{2}\right)=: B
$$

This we get as follows:

$$
\begin{aligned}
B-A & =\sum_{i=1}^{n} x_{i}^{2} \sum_{j=1}^{n} y_{j}^{2}-\sum_{i=1}^{n} x_{i} y_{i} \sum_{j=1}^{n} x_{j} y_{j} \\
& =\frac{1}{2} \sum_{i=1}^{n} x_{i}^{2} \sum_{j=1}^{n} y_{j}^{2}+\frac{1}{2} \sum_{j=1}^{n} x_{j}^{2} \sum_{i=1}^{n} y_{i}^{2}-\sum_{i=1}^{n} x_{i} y_{i} \sum_{j=1}^{n} x_{j} y_{j} \\
& =\sum_{i, j=1}^{n} \frac{1}{2}\left(x_{i}^{2} y_{j}^{2}+x_{j}^{2} y_{i}^{2}-2 x_{i} y_{i} x_{j} y_{j}\right) \\
& =\sum_{i, j=1}^{n} \frac{1}{2}\left(x_{i} y_{j}-x_{j} y_{i}\right)^{2} \\
& \geq 0 .
\end{aligned}
$$

An inner product on $X$ always induces a norm on $X$, which is defined by

$$
\|x\|=\langle x, x\rangle^{\frac{1}{2}}
$$

for every $x \in X$. To show that $\|$.$\| is a norm on X$ we need the inequality

$$
|\langle x, y\rangle| \leq\|x\|\|y\|,
$$

which generalizes the inequality of Cauchy.

Definition 1.7.6. A metric $d$ on a set $X$ is a function $d: X \times X \rightarrow \mathbb{R}$ such that for every $x, y, z \in X$ the following hold:
(i) $d(x, y) \geq 0$.
(ii) $d(x, y)=0 \Leftrightarrow x=y$.
(iii) $d(x, y)=d(y, x)$.
(iv) $d(x, y) \leq d(x, z)+d(z, y)$.

If $d$ is a metric on $X$, the pair $(X, d)$ is called a metric space.
A norm $\|$.$\| on a linear space X$ induces a metric on $X$ defined by

$$
d(x, y):=\|x-y\| .
$$

Definition 1.7.7. The Euclidean metric $d$ on $\mathbb{R}^{n}$ is the metric induced by the Euclidean norm on $\mathbb{R}^{n}$ i.e.,

$$
\begin{gathered}
d(x, y):=|x-y|:=\left(\sum_{i=1}^{n}\left(x_{i}-y_{i}\right)^{2}\right)^{\frac{1}{2}}= \\
=\sqrt{\left(x_{1}-y_{1}\right)^{2}+\left(x_{2}-y_{2}\right)^{2}+\ldots+\left(x_{n}-y_{n}\right)^{2}}=\sqrt{\langle x-y, x-y\rangle},
\end{gathered}
$$

for every $x, y \in \mathbb{R}^{n}$.
The Euclidean norm is the norm induced by the Euclidean inner product. To understand the geometric meaning of the Euclidean inner product we first see that a vector $x \in \mathbb{R}^{n}$ is orthogonal to a vector $y \in \mathbb{R}^{n}$, in symbols $x \perp y$, if and only if $\langle x, y\rangle=0$. To explain this we work as follows. It is easy to see geometrically ${ }^{6}$ that

$$
x \perp y \Leftrightarrow|x-y|=|x+y|
$$

since the diagonals of the parallelogram are equal only if $x$ is perpendicular to $y$.


[^5]We show that

$$
|x-y|=|x+y| \Leftrightarrow\langle x, y\rangle=0 .
$$

Since $|x| \geq 0$, we have that

$$
\begin{aligned}
|x-y|=|x+y| & \Leftrightarrow|x-y|^{2}=|x+y|^{2} \\
& : \Leftrightarrow\langle x-y, x-y\rangle=\langle x+y, x+y\rangle \\
& \Leftrightarrow\langle x, x\rangle-2\langle x, y\rangle+\langle y, y\rangle=\langle x, x\rangle+2\langle x, y\rangle+\langle y, y\rangle \\
& \Leftrightarrow 4\langle x, y\rangle=0 \\
& \Leftrightarrow\langle x, y\rangle=0 .
\end{aligned}
$$

By the last two equivalences we get the required equivalence

$$
x \perp y \Leftrightarrow\langle x, y\rangle=0
$$

Corollary 1.7.8 (Pythagoras theorem). If $x, y \in \mathbb{R}^{n}$, such that $x \perp y$, then

$$
|x+y|^{2}=|x|^{2}+|y|^{2}
$$

Proof. Exercise.
By the inequality of Cauchy we have that

$$
\left|\frac{|\langle x, y\rangle|}{|x||y|}\right|=\frac{|\langle x, y\rangle|}{|x||y|} \leq 1 \Leftrightarrow-1 \leq \frac{\langle x, y\rangle}{|x||y|} \leq 1
$$

hence, there exists a unique angle $\theta \in[0, \pi]$ such that

$$
\cos \theta=\frac{\langle x, y\rangle}{|x||y|}
$$

and we call $\theta$ the angle between $x$ and $y$. Clearly, if $\langle x, y\rangle=0$, then $\theta=\frac{\pi}{2}$.
Proposition 1.7.9. If $x, y \in \mathbb{R}^{n}$, and $y \neq \mathbf{0}$, then the projection $\operatorname{pr}_{y}(x)$ of $x$ on $y$ is given by


$$
\operatorname{pr}_{y}(x):=\lambda y \quad \& \quad \lambda:=\frac{\langle x, y\rangle}{\langle y, y\rangle} .
$$

Proof. Since $(x-\lambda y) \perp y$, and $y \neq \mathbf{0}$, we have that

$$
\begin{aligned}
\langle(x-\lambda y), y\rangle=0 & \Leftrightarrow\langle x, y\rangle-\langle\lambda y, y\rangle=0 \\
& \Leftrightarrow\langle x, y\rangle-\lambda\langle y, y\rangle=0 \\
& \Leftrightarrow \lambda=\frac{\langle x, y\rangle}{\langle y, y\rangle} .
\end{aligned}
$$

## CHAPTER 2

## Functions of several variables

### 2.1. Curves in $\mathbb{R}^{n}$

Definition 2.1.1. Let $I$ be an interval of $\mathbb{R}$ of the form

$$
(-\infty, a),(-\infty, a],(a,+\infty),[a,+\infty), \mathbb{R},(a, b),(a, b],[a, b),[a, b]
$$

where $a, b \in \mathbb{R}$ such that $a \leq b$. A curve in $\mathbb{R}^{n}$ is a function

$$
\boldsymbol{x}: I \rightarrow \mathbb{R}^{n} \quad I \ni t \mapsto \boldsymbol{x}(t) \in \mathbb{R}^{n}, \quad t \in I .
$$

We also write

$$
\boldsymbol{x}(t)=\left(x_{1}(t), \ldots, x_{n}(t)\right), \quad t \in I
$$

where $x_{i}: I \rightarrow \mathbb{R}$ is the $i$-coordinate function of $\boldsymbol{x}$, for every $i \in\{1, \ldots, n\}$. We also call $\boldsymbol{x}(t)$ the position vector of $\boldsymbol{x}$ at time $t$. We call $\boldsymbol{x}$ differentiable on (every element of) $I$, if the coordinate functions $x_{1}(t), \ldots, x_{n}(t)$ of $\boldsymbol{x}$ are differentiable on (every element of) $I$. A point $P \in \mathbb{R}^{n}$ belongs to $\boldsymbol{x}$, if there is some $t \in I$ such that $P=\boldsymbol{x}(t)$.

Next we draw the image of a differentiable curve $\boldsymbol{x}:[a, b] \rightarrow \mathbb{R}^{2}$

and the image of a differentiable curve $\boldsymbol{x}:(a, b) \rightarrow \mathbb{R}^{2}$.


Let also the curve $\boldsymbol{c}:[0,2 \pi] \rightarrow \mathbb{R}^{2}$, defined by $\theta \mapsto(\cos \theta, \sin \theta)$, for every $\theta \in[0,2 \pi]$, the image of which is the unit circle in $\mathbb{R}^{2}$.


This is a differentiable curve, since $\boldsymbol{c}(\theta):=\left(c_{1}(\theta), c_{2}(\theta)\right)$, and its coordinate functions $c_{1}(\theta):=\cos \theta$, and $c_{2}(\theta):=\sin \theta$ are differentiable on $[0,2 \pi]$, since $\cos ^{\prime} \theta=$ $-\sin \theta$, and $\sin ^{\prime} \theta=\cos \theta$, for every $\theta \in[0,2 \pi]$. Moreover, $\boldsymbol{c}$ is a closed curve, since $\boldsymbol{c}(0)=\boldsymbol{c}(2 \pi)$.

If $\boldsymbol{x}(t): I \rightarrow \mathbb{R}^{n}$ is a differentiable curve in $\mathbb{R}^{n}, t_{0} \in I$, and $h \in \mathbb{R}$, then

$$
\begin{aligned}
\frac{\boldsymbol{x}\left(t_{0}+h\right)-\boldsymbol{x}\left(t_{0}\right)}{h} & =\frac{1}{h}\left[\left(x_{1}\left(t_{0}+h\right), \ldots, x_{n}\left(t_{0}+h\right)\right)-\left(x_{1}\left(t_{0}\right), \ldots, x_{n}\left(t_{0}\right)\right)\right] \\
& =\frac{1}{h}\left(x_{1}\left(t_{0}+h\right)-x_{1}\left(t_{0}\right), \ldots, x_{n}\left(t_{0}+h\right)-x_{n}\left(t_{0}\right)\right) \\
& =\left(\frac{x_{1}\left(t_{0}+h\right)-x_{1}\left(t_{0}\right)}{h}, \ldots, \frac{x_{n}\left(t_{0}+h\right)-x_{n}\left(t_{0}\right)}{h}\right)
\end{aligned}
$$

and hence

$$
\lim _{h \rightarrow 0} \frac{\boldsymbol{x}\left(t_{0}+h\right)-\boldsymbol{x}\left(t_{0}\right)}{h}=\left(x_{1}^{\prime}\left(t_{0}\right), \ldots, x_{n}^{\prime}\left(t_{0}\right)\right) .
$$

Definition 2.1.2. If $\boldsymbol{x}: I \rightarrow \mathbb{R}^{n}$ is a differentiable curve, its derivative is the curve $\boldsymbol{x}^{\prime}: I \rightarrow \mathbb{R}^{n}$ defined, for every $t_{0} \in I$, by

$$
\boldsymbol{x}^{\prime}\left(t_{0}\right):=\frac{d \boldsymbol{x}}{d t}\left(t_{0}\right):=\left(x_{1}^{\prime}\left(t_{0}\right), \ldots, x_{n}^{\prime}\left(t_{0}\right)\right):=\left(\frac{d x_{1}}{d t}\left(t_{0}\right), \ldots, \frac{d x_{n}}{d t}\left(t_{0}\right)\right)
$$

We call $\boldsymbol{x}^{\prime}\left(t_{0}\right)$ the velocity vector of $\boldsymbol{x}(t)$ at time $t_{0}$.
The velocity vector $\boldsymbol{x}^{\prime}\left(t_{0}\right)$ is located at the origin of the Euclidean plane, but we view it as a vector tangent to the curve at $t_{0}$ and parallel to it.


Definition 2.1.3. Let $\boldsymbol{x}: I \rightarrow \mathbb{R}^{n}$ be a differentiable curve. Its speed $v_{\boldsymbol{x}}: I \rightarrow$ $[0,+\infty)$ is defined, for every $t \in I$, by

$$
v_{\boldsymbol{x}}(t):=\left|\boldsymbol{x}^{\prime}(t)\right|,
$$

where $\left|\boldsymbol{x}^{\prime}(t)\right|$ is the Euclidean norm of the vector $\boldsymbol{x}^{\prime}(t)$. If the derivative $\boldsymbol{x}^{\prime}: I \rightarrow \mathbb{R}^{n}$ of $\boldsymbol{x}$ is differentiable, the acceleration vector of $\boldsymbol{x}(t)$ at time $t_{0} \in I$ is defined by

$$
\boldsymbol{x}^{\prime \prime}\left(t_{0}\right):=\frac{d \boldsymbol{x}^{\prime}}{d t}\left(t_{0}\right):=\frac{d^{2} \boldsymbol{x}}{d t}\left(t_{0}\right) .
$$

Notice that by the definition of the Euclidean norm |.| we have that

$$
v_{\boldsymbol{x}}(t)^{2}:=\left|\boldsymbol{x}^{\prime}(t)\right|^{2}=\left\langle x^{\prime}(t), x^{\prime}(t)\right\rangle .
$$

Proposition 2.1.4. Let $\boldsymbol{x}, \boldsymbol{y}: I \rightarrow \mathbb{R}^{n}$ be differentiable curves, $\lambda \in \mathbb{R}$, and $f: I \rightarrow \mathbb{R}$ a differentiable function.
(i) The sum $\boldsymbol{x}+\boldsymbol{y}: I \rightarrow \mathbb{R}^{n}$, defined by

$$
(\boldsymbol{x}+\boldsymbol{y})(t):=\boldsymbol{x}(t)+\boldsymbol{y}(t)
$$

for every $t \in I$, is a differentiable curve, and, for every $t_{0} \in I$, we have that

$$
(\boldsymbol{x}+\boldsymbol{y})^{\prime}\left(t_{0}\right)=\boldsymbol{x}^{\prime}\left(t_{0}\right)+\boldsymbol{y}^{\prime}\left(t_{0}\right)
$$

(ii) The product $\lambda \boldsymbol{x}: I \rightarrow \mathbb{R}^{n}$, defined by

$$
(\lambda \boldsymbol{x})(t):=\lambda \boldsymbol{x}(t),
$$

for every $t \in I$, is a differentiable curve, and, for every $t_{0} \in I$, we have that

$$
(\lambda \boldsymbol{x})^{\prime}\left(t_{0}\right)=\lambda \boldsymbol{x}^{\prime}\left(t_{0}\right) .
$$

(iii) The product $\langle\boldsymbol{x}, \boldsymbol{y}\rangle: I \rightarrow \mathbb{R}$, defined by

$$
\langle\boldsymbol{x}, \boldsymbol{y}\rangle(t):=\langle\boldsymbol{x}(t), \boldsymbol{y}(t)\rangle,
$$

for every $t \in I$, where $\langle\boldsymbol{x}(t), \boldsymbol{y}(t)\rangle$ is the Euclidean inner product of $\boldsymbol{x}(t), \boldsymbol{y}(t)$, is a differentiable function, and, for every $t_{0} \in I$, we have that

$$
\langle\boldsymbol{x}, \boldsymbol{y}\rangle^{\prime}\left(t_{0}\right)=\left\langle\boldsymbol{x}^{\prime}\left(t_{0}\right), \boldsymbol{y}\left(t_{0}\right)\right\rangle+\left\langle\boldsymbol{x}\left(t_{0}\right), \boldsymbol{y}^{\prime}\left(t_{0}\right)\right\rangle .
$$

(iv) The product $\boldsymbol{x}^{2}: I \rightarrow \mathbb{R}$, defined by

$$
\left(\boldsymbol{x}^{2}\right)(t):=\langle\boldsymbol{x}(t), \boldsymbol{x}(t)\rangle,
$$

for every $t \in I$, is a differentiable function, and, for every $t_{0} \in I$, we have that

$$
\left(\boldsymbol{x}^{2}\right)^{\prime}\left(t_{0}\right)=2\left\langle\boldsymbol{x}\left(t_{0}\right), \boldsymbol{x}^{\prime}\left(t_{0}\right)\right\rangle
$$

(v) The product $f \boldsymbol{x}: I \rightarrow \mathbb{R}^{n}$, defined by

$$
(f \boldsymbol{x})(t):=f(t) \boldsymbol{x}(t),
$$

for every $t \in I$, is a differentiable curve, and, for every $t_{0} \in I$, we have that

$$
(f \boldsymbol{x})^{\prime}\left(t_{0}\right)=f^{\prime}\left(t_{0}\right) \boldsymbol{x}\left(t_{0}\right)+f\left(t_{0}\right) \boldsymbol{x}^{\prime}\left(t_{0}\right)
$$

Proof. We prove only the case (iii), and the rest is an exercise. By the definition of the Euclidean inner product we have that

$$
\langle\boldsymbol{x}, \boldsymbol{y}\rangle(t):=\langle\boldsymbol{x}(t), \boldsymbol{y}(t)\rangle:=\sum_{i=1}^{n} x_{i}(t) y_{i}(t)=x_{1}(t) y_{1}(t)+\ldots+x_{n}(t) y_{n}(t),
$$

hence we have that

$$
\begin{aligned}
& \langle\boldsymbol{x}, \boldsymbol{y}\rangle^{\prime}\left(t_{0}\right)= \\
& =\left[x_{1}(t) y_{1}(t)\right]^{\prime}\left(t_{0}\right)+\ldots+\left[x_{n}(t) y_{n}(t)\right]^{\prime}\left(t_{0}\right) \\
& =\left[x_{1}^{\prime}\left(t_{0}\right) y_{1}\left(t_{0}\right)+x_{1}\left(t_{0}\right) y_{1}^{\prime}\left(t_{0}\right)\right]+\ldots+\left[x_{n}^{\prime}\left(t_{0}\right) y_{n}\left(t_{0}\right)+x_{n}\left(t_{0}\right) y_{n}^{\prime}\left(t_{0}\right)\right] \\
& =\left[x_{1}^{\prime}\left(t_{0}\right) y_{1}\left(t_{0}\right)+\ldots+x_{n}^{\prime}\left(t_{0}\right) y_{n}\left(t_{0}\right)\right]+\left[x_{1}\left(t_{0}\right) y_{1}^{\prime}\left(t_{0}\right)+\ldots+x_{n}\left(t_{0}\right) y_{n}^{\prime}\left(t_{0}\right)\right] \\
& =\sum_{i=1}^{n} x_{i}^{\prime}\left(t_{0}\right) y_{i}\left(t_{0}\right)+\sum_{i=1}^{n} x\left(t_{0}\right) y_{i}^{\prime}\left(t_{0}\right) \\
& :=\left\langle\boldsymbol{x}^{\prime}\left(t_{0}\right), \boldsymbol{y}\left(t_{0}\right)\right\rangle+\left\langle\boldsymbol{x}\left(t_{0}\right), \boldsymbol{y}^{\prime}\left(t_{0}\right)\right\rangle .
\end{aligned}
$$

Corollary 2.1.5. Let $\boldsymbol{x}: I \rightarrow \mathbb{R}^{n}$ be a differentiable curve such that for every $t \in I$ the distance of $\boldsymbol{x}(t)$ from the origin remains constant i.e.,

$$
|\boldsymbol{x}(t)|=r>0
$$

for every $t \in I$. Then for every $t_{0} \in I$ the position vector $\boldsymbol{x}\left(t_{0}\right)$ of $\boldsymbol{x}$ at $t_{0}$ is orthogonal to the velocity vector $\boldsymbol{x}^{\prime}\left(t_{0}\right)$ of $\boldsymbol{x}$ at $t_{0}$.


Proof. If $|\boldsymbol{x}(t)|=r>0$, for every $t \in I$, then $\boldsymbol{x}(t)$ lies on the sphere of radius r. Moreover,

$$
r^{2}=|\boldsymbol{x}(t)|^{2}=\langle\boldsymbol{x}(t), \boldsymbol{x}(t)\rangle:=\langle\boldsymbol{x}, \boldsymbol{x}\rangle(t),
$$

hence by the Proposition 2.1.4(iv), and since $\langle\boldsymbol{x}, \boldsymbol{x}\rangle$ is a constant function on $I$, we have that

$$
0=\langle\boldsymbol{x}, \boldsymbol{x}\rangle^{\prime}\left(t_{0}\right)=2\left\langle\boldsymbol{x}\left(t_{0}\right), \boldsymbol{x}^{\prime}\left(t_{0}\right)\right\rangle \Leftrightarrow 0=\left\langle\boldsymbol{x}\left(t_{0}\right), \boldsymbol{x}^{\prime}\left(t_{0}\right)\right\rangle \Leftrightarrow \boldsymbol{x}\left(t_{0}\right) \perp \boldsymbol{x}^{\prime}\left(t_{0}\right) .
$$

Definition 2.1.6. If $\boldsymbol{x}: I \rightarrow \mathbb{R}$ is a differentiable curve with continuous derivative $\boldsymbol{x}^{\prime}$, its length $L_{a b}(\boldsymbol{x})$ between two values $a, b \in I$, where $a \leq b$, is defined by the corresponding integral of its speed i.e.,

$$
L_{a, b}(\boldsymbol{x}):=\int_{a}^{b} v_{\boldsymbol{x}}(t) d t:=\int_{a}^{b}\left|\boldsymbol{x}^{\prime}(t)\right| d t
$$

By the definition of the Euclidean norm we have that

$$
L_{a, b}(\boldsymbol{x})=\int_{a}^{b} \sqrt{\left(\frac{d x_{1}}{d t}(t)\right)^{2}+\left(\frac{d x_{2}}{d t}(t)\right)^{2}} d t
$$

if $\boldsymbol{x}(t):=\left(x_{1}(t), x_{2}(t)\right)$, and

$$
L_{a, b}(\boldsymbol{x})=\int_{a}^{b} \sqrt{\left(\frac{d x_{1}}{d t}(t)\right)^{2}+\left(\frac{d x_{2}}{d t}(t)\right)^{2}+\left(\frac{d x_{3}}{d t}(t)\right)^{2}} d t
$$

if $\boldsymbol{x}(t):=\left(x_{1}(t), x_{2}(t), x_{3}(t)\right)$. In the general case, where $\boldsymbol{x}(t):=\left(x_{1}(t), \ldots, x_{n}(t)\right)$, we have that

$$
L_{a, b}(\boldsymbol{x})=\int_{a}^{b} \sqrt{\left(\frac{d x_{1}}{d t}(t)\right)^{2}+\ldots+\left(\frac{d x_{3}}{d t}(t)\right)^{2}} d t
$$

If for example, we consider the unit circle $\boldsymbol{c}(\theta):=(\cos \theta, \sin \theta)$, where $\theta \in[0,2 \pi]$, then we have that

$$
\begin{aligned}
v_{\boldsymbol{c}}(\theta) & :=\left|\boldsymbol{c}^{\prime}(\theta)\right| \\
& :=\sqrt{c_{1}^{\prime}(\theta)^{2}+c_{2}^{\prime}(\theta)^{2}} \\
& =\sqrt{(-\sin \theta)^{2}+(\cos \theta)^{2}} \\
& =\sqrt{\sin ^{2} \theta+\cos ^{2} \theta} \\
& =\sqrt{1} \\
& =1
\end{aligned}
$$

and hence we get the expected value for the length of $\boldsymbol{c}$ between 0 and $2 \pi$ :

$$
L_{0,2 \pi}(\boldsymbol{c}):=\int_{0}^{2 \pi} v_{\boldsymbol{c}}(\theta) d \theta:=\int_{0}^{2 \pi} 1 d \theta=\int_{0}^{2 \pi} d \theta=2 \pi-0=2 \pi .
$$

Let the differentiable curve $\boldsymbol{x}: \mathbb{R} \rightarrow \mathbb{R}^{2}$ defined by

$$
\boldsymbol{x}(t):=\left(e^{t} \cos t, e^{t} \sin t\right),
$$

for every $t \in \mathbb{R}$. Its derivative $\boldsymbol{x}^{\prime}$ is given by

$$
\boldsymbol{x}^{\prime}(t):=\left(e^{t} \cos t-e^{t} \sin t, e^{t} \sin t+e^{t} \cos t\right)
$$

for every $t \in \mathbb{R}$. After some calculations we get

$$
|\boldsymbol{x}(t)|=e^{t} \quad \& \quad\left|\boldsymbol{x}^{\prime}(t)\right|=\sqrt{2} e^{t} \quad \& \quad\left\langle\boldsymbol{x}^{\prime}(t), \boldsymbol{x}(t)\right\rangle=e^{2 t}
$$

for every $t \in \mathbb{R}$. Hence,

$$
\frac{\left\langle\boldsymbol{x}^{\prime}(t), \boldsymbol{x}(t)\right\rangle}{\left|\boldsymbol{x}^{\prime}(t)\right||\boldsymbol{x}(t)|}=\frac{e^{2 t}}{\sqrt{2} e^{t} e^{t}}=\frac{1}{\sqrt{2}},
$$

for every $t \in \mathbb{R}$ i.e., the angle between $\boldsymbol{x}^{\prime}(t)$ and $\boldsymbol{x}(t)$ is constant $\frac{\pi}{4}$, for every $t \in \mathbb{R}$. Moreover,

$$
L_{0,1}(\boldsymbol{x})=\int_{0}^{1} \sqrt{2} e^{t} d t=\sqrt{2}(e-1)
$$

### 2.2. Open sets in $\mathbb{R}^{n}$

We consider vector-valued functions defined on appropriate subsets of $\mathbb{R}^{n}$ that we call open.

Definition 2.2.1. Let $x \in \mathbb{R}^{n}$ and $\epsilon>0$. The open ball $\mathcal{B}(x, \epsilon)$ with center $x$ and radius $\epsilon$ is defined by

$$
\begin{aligned}
\mathcal{B}(x, \epsilon) & :=\left\{y \in \mathbb{R}^{n} \mid d(x, y)<\epsilon\right\} \\
& :=\left\{y \in \mathbb{R}^{n}| | x-y \mid<\epsilon\right\} \\
& :=\left\{y \in \mathbb{R}^{n} \mid \sqrt{\left(x_{1}-y_{1}\right)^{2}+\ldots\left(x_{n}-y_{n}\right)^{2}}<\epsilon\right\} .
\end{aligned}
$$

We also say that $\mathcal{B}(x, \epsilon)$ is the open $r$-ball at $x$. The closed ball $\mathcal{B}(x, \epsilon]$ with center $x$ and radius $\epsilon$ is defined by

$$
\mathcal{B}(x, \epsilon]:=\left\{y \in \mathbb{R}^{n} \mid d(x, y) \leq \epsilon\right\} .
$$

If $U \subseteq \mathbb{R}^{n}$, we say that $U$ is an open subset of $\mathbb{R}^{n}$, if

$$
\forall_{x \in U} \exists_{\epsilon>0}(\mathcal{B}(x, \epsilon) \subseteq U)
$$

If $F \subseteq \mathbb{R}^{n}$, we say that $F$ is a closed subset of $\mathbb{R}^{n}$, if its complement

$$
F^{c}:=\left\{y \in \mathbb{R}^{n} \mid x \notin F\right\}
$$

is open.
The open $\epsilon$-ball $\mathcal{B}(\mathbf{0}, \epsilon)$ at the origin $(0,0)$ is the open $\epsilon$-disc around $(0,0)$

and the closed $\epsilon$-ball $\mathcal{B}(\mathbf{0}, \epsilon]$ at the origin $(0,0)$ is the $\epsilon$-disc around $(0,0)$ with the $\epsilon$-circle around the origin. It is easy to see that the open $\epsilon$-ball $\mathcal{B}(\mathbf{0}, \epsilon)$, as any open ball, is an open set, since if we take any point in the disc, we can find a small disc around it that is included in the larger one. Note that the closed $\epsilon$-ball $\mathcal{B}(\mathbf{0}, \epsilon]$ is not open, since any disc around a point at the $\epsilon$-circle is not included in $\mathcal{B}(\mathbf{0}, \epsilon]$. It
is clear though, that $\mathcal{B}(\mathbf{0}, \epsilon]$ is closed. Using a similar argument we can show that the interior $U$ of the following curve in $\mathbb{R}^{2}$ is open in $\mathbb{R}^{2}$.


Note that the open $\epsilon$-ball $\mathcal{B}(\mathbf{0}, \epsilon)$ in $\mathbb{R}$ at the origin 0 is the open interval $(-\epsilon, \epsilon)$.
Proposition 2.2.2. Let $n \geq 1$.
(i) $\mathbb{R}^{n}$ and $\emptyset$ are both open and closed.
(ii) If $U \subseteq \mathbb{R}^{n}$, then $U$ is open if and only if its complement $U^{c}$ is closed.
(iii) If $U, V$ are open in $\mathbb{R}^{n}$, then $U \cap V$ and $U \cup V$ are open in $\mathbb{R}^{n}$.
(iv) If $F, K$ are closed in $\mathbb{R}^{n}$, then $F \cap K$ and $F \cup K$ are closed in $\mathbb{R}^{n}$.
(v) If $\left(U_{i}\right)_{i \in I}$ is a family of open sets in $\mathbb{R}^{n}$ i.e., $U_{i}$ is open for every $i \in I$, then their union

$$
\bigcup_{i \in I} U_{i}:=\left\{x \in \mathbb{R}^{n} \mid \exists_{i \in I}\left(x \in U_{i}\right)\right\}
$$

is open.
(vi) If $\left(F_{i}\right)_{i \in I}$ is a family of closed sets in $\mathbb{R}^{n}$ i.e., $U_{i}$ is closed for every $i \in I$, then their intersection

$$
\bigcap_{i \in I} F_{i}:=\left\{x \in \mathbb{R}^{n} \mid \forall_{i \in I}\left(x \in F_{i}\right)\right\}
$$

is closed.

Proof. (i) If $x \in \mathbb{R}^{n}$, then $\mathcal{B}(x, 1) \subseteq \mathbb{R}^{n}$, and hence $\mathbb{R}^{n}$ is open. Consequently, $\emptyset$ is closed, since $\emptyset^{c}=\mathbb{R}^{n}$. The implication $x \in \emptyset \Rightarrow \mathcal{B}(x, 1) \subseteq \emptyset$ is trivially true, since its premise is false. Hence $\emptyset$ is also open, and $\mathbb{R}^{n}$ is also closed, since $\left(\mathbb{R}^{n}\right)^{c}=\emptyset$.
(ii) If $U$ is open, then $U^{c}$ is closed, since $\left(U^{c}\right)^{c}=U$ is open. If $U^{c}$ is closed, then by definition $\left(U^{c}\right)^{c}=U$ is open.
(iii) First we show that $U \cap V$ is open. If $x \in U \cap V$, then $x \in U$ and $x \in V$. Since $U$ is open, there is some $\epsilon_{1}>0$ such that $\mathcal{B}\left(x, \epsilon_{1}\right) \subseteq U$. Since $V$ is open, there is some $\epsilon_{2}>0$ such that $\mathcal{B}\left(x, \epsilon_{2}\right) \subseteq Y$. If

$$
\epsilon:=\min \left\{\epsilon_{1}, \epsilon_{2}\right\}
$$

then

$$
\mathcal{B}(x, \epsilon) \subseteq V \cap U
$$

To show this, let $y \in \mathbb{R}^{n}$ such that $|y-x|<\epsilon \leq \epsilon_{1}$. Hence $y \in U$. Similarly, $|y-x|<\epsilon \leq \epsilon_{2}$, and hence $y \in Y$. Consequently, $y \in V \cap U$.

Next we show that $U \cup V$ is open. If $x \in U \cup V$, then $x \in U$, or $x \in V$. In the first case we have that $\mathcal{B}\left(x, \epsilon_{1}\right) \subseteq U \subseteq U \cup V$, and in the second we have that $\mathcal{B}\left(x, \epsilon_{2}\right) \subseteq V \subseteq U \cup V$.
(iv) We use the case (iii) and the equalities

$$
(F \cap K)^{c}=F^{c} \cup K^{c} \quad \& \quad(F \cup K)^{c}=F^{c} \cap K^{c} .
$$

(v) and (vi) is an exercise.

The intersection of a countable family of open sets is not generally open. E.g.,

$$
(0,1]=\bigcap_{n \geq 1}\left(0,1+\frac{1}{n}\right)
$$

and $(0,1]$ is not open, as any non-trivial interval around 1 intersects $(1,+\infty)$. The union of a countable family of closed sets is not generally closed. E.g.,

$$
(0,1)=\bigcup_{n \geq 2}\left[\frac{1}{n}, 1-\frac{1}{n}\right]
$$

and $(0,1)$ is not closed, since its complement $(-\infty, 0] \cup[1,+\infty)$ is not open. It is not hard to see that the cartesian product of open sets in $\mathbb{R}$ is an open set in the corresponding $\mathbb{R}^{n}$. E.g., the set

$$
(0,1) \times(-1,1):=\left\{(x, y) \in \mathbb{R}^{2} \mid x \in(0,1) \& y \in(-1,1)\right\}
$$

is open in $\mathbb{R}^{2}$. Similarly the set

$$
(0,1) \times(-1,1) \times \mathbb{R}:=\left\{(x, y, z) \in \mathbb{R}^{3} \mid x \in(0,1) \& y \in(-1,1)\right\}
$$

is open in $\mathbb{R}^{3}$.

### 2.3. Partial derivatives

If $U$ is an open subset of $\mathbb{R}^{n}$, and $x=\left(x_{1}, \ldots, x_{n}\right) \in U$, then for every $i \in$ $\{1, \ldots, n\}$, there are appropriately small values of $h \in \mathbb{R}$ such that the point

$$
\left(x_{1}, \ldots, x_{i}+h, \ldots, x_{n}\right) \in U
$$

and the following concept is well-defined.

Definition 2.3.1. Let $U$ be an open subset of $\mathbb{R}^{n}, x=\left(x_{1}, \ldots, x_{n}\right) \in U$, and $f: U \rightarrow \mathbb{R}$. If the following limit exists

$$
\lim _{h \rightarrow 0} \frac{f\left(x_{1}, \ldots, x_{i}+h, \ldots, x_{n}\right)-f\left(x_{1}, \ldots, x_{n}\right)}{h}
$$

we let

$$
D_{i} f(x):=\frac{\partial f}{\partial x_{i}}(x):=\lim _{h \rightarrow 0} \frac{f\left(x_{1}, \ldots, x_{i}+h, \ldots, x_{n}\right)-f\left(x_{1}, \ldots, x_{n}\right)}{h}
$$

and we call $D_{i} f(x)$, or $\frac{\partial f}{\partial x_{i}}(x)$, the $i$-th partial derivative of $f$ at $x$.
If $B_{n}:=\left\{e_{1}, \ldots, e_{i}, \ldots, e_{n}\right\}$ is the standard basis of $\mathbb{R}^{n}$, we have that

$$
D_{i} f(x)=\lim _{h \rightarrow 0} \frac{f\left(x+h e_{i}\right)-f(x)}{h}
$$

If for example $f: \mathbb{R}^{2} \rightarrow \mathbb{R}$ is defined by

$$
f(x, y):=x^{2} y^{3}
$$

then

$$
\begin{aligned}
D_{1} f(x) & :=\frac{\partial f}{\partial x}(x) \\
& :=\lim _{h \rightarrow 0} \frac{f(x+h, y)-f(x, y)}{h} \\
& =\lim _{h \rightarrow 0} \frac{(x+h)^{2} y^{3}-x^{2} y^{3}}{h} \\
& =y^{3} \lim _{h \rightarrow 0} \frac{(x+h)^{2}-x^{2}}{h} \\
& =y^{3} 2 x \\
& =2 x y^{3}
\end{aligned}
$$

since the term in the right is the derivative of the function $g(x)=x^{2}$. I.e., to calculate $D_{1} f(x)$ we treat $y$ as a constant and we differentiate with respect to $x$. Similarly we have that

$$
\begin{aligned}
D_{2} f(x) & :=\frac{\partial f}{\partial y}(x) \\
& :=\lim _{h \rightarrow 0} \frac{f(x, y+h)-f(x, y)}{h} \\
& =\lim _{h \rightarrow 0} \frac{x^{2}(y+h)^{3}-x^{2} y^{3}}{h} \\
& =x^{2} \lim _{h \rightarrow 0} \frac{(y+h)^{3}-y^{3}}{h} \\
& =x^{2} 3 y^{2} \\
& =3 x^{2} y^{2}
\end{aligned}
$$

since the term in the right is the derivative of the function $h(y)=y^{3}$. I.e., to calculate $D_{2} f(x)$ we treat $x$ as a constant and we differentiate with respect to $y$.

If $f, g: U \rightarrow \mathbb{R}$, and $x \in U$ such that $D_{i} f(x)$ and $D_{i} g(x)$ exist, then by the properties of the derivative of real-valued functions on intervals of $\mathbb{R}$ we get immediately

$$
\begin{gathered}
D_{i}(f+g)(x)=D_{i} f(x)+D_{i} g(x) \\
D_{i}(\lambda f)(x)=\lambda D_{i} f(x)
\end{gathered}
$$

for every $\lambda \in \mathbb{R}$.
Definition 2.3.2. Let $U$ be an open subset of $\mathbb{R}^{n}, x ;=\left(x_{1}, \ldots, x_{n}\right) \in U$, and $f: U \rightarrow \mathbb{R}$. If the partial derivatives at $x$

$$
D_{1} f(x):=\frac{\partial f}{\partial x_{1}}(x), \ldots, D_{n} f(x):=\frac{\partial f}{\partial x_{n}}(x)
$$

exist, the gradient $(\operatorname{grad} f)(x)$ of $f$ at $x$ is the vector

$$
\begin{aligned}
(\operatorname{grad} f)(x) & :=\left(\frac{\partial f}{\partial x_{1}}(x), \ldots, \frac{\partial f}{\partial x_{n}}(x)\right) \\
& :=\left(D_{1} f(x), \ldots, D_{n} f(x)\right)
\end{aligned}
$$

E.g., if $f: \mathbb{R}^{2} \rightarrow \mathbb{R}$ is defined as above by $f(x, y):=x^{2} y^{3}$, then

$$
(\operatorname{grad} f)(x):=\left(2 x y^{3}, 3 x^{2} y^{2}\right)
$$

Because of the above linearity of $D_{i}$, we get immediately that if $f, g: U \rightarrow \mathbb{R}$, and $x \in U$ such that $D_{i} f(x)$ and $D_{i} g(x)$ exist, then

$$
\begin{gathered}
(\operatorname{grad}(f+g))(x)=(\operatorname{grad} f)(x)+(\operatorname{grad} g)(x), \\
(\operatorname{grad}(\lambda f))(x)=\lambda(\operatorname{grad} f)(x),
\end{gathered}
$$

for every $\lambda \in \mathbb{R}$. If $D_{i} f(x)$ and $D_{i} g(x)$ exist, for every $x \in U$, we get

$$
\begin{gathered}
\operatorname{grad}(f+g)=\operatorname{grad} f+\operatorname{grad} g \\
\operatorname{grad}(\lambda f)=\lambda \operatorname{grad} f
\end{gathered}
$$

for every $\lambda \in \mathbb{R}$. If $f: \mathbb{R}^{2} \rightarrow \mathbb{R}$ is defined by $f(x, y):=x^{2} y^{3}$, we showed that

$$
D_{1} f(x):=\frac{\partial f}{\partial x}(x)=2 x y^{3} \quad \& \quad D_{2} f(x):=\frac{\partial f}{\partial y}(x)=3 x^{2} y^{2} .
$$

Since $D_{1} f, D_{2} f: \mathbb{R}^{2} \rightarrow \mathbb{R}$, we can determine the repeated partial derivatives

$$
\begin{gathered}
D_{1} D_{1} f(x, y):=D_{1}^{2} f(x, y):=\frac{\partial^{2} f}{\partial x^{2}}(x, y):=\left(D_{1}\left(D_{1} f\right)\right)(x, y)= \\
=\frac{\partial\left(2 x y^{3}\right)}{\partial x}(x, y)=2 y^{3}, \\
D_{1} D_{2} f(x, y):=\frac{\partial}{\partial x}\left(\frac{\partial f}{\partial y}\right)(x, y):=\frac{\partial^{2} f}{\partial x \partial y}(x, y):=\left(D_{1}\left(D_{2} f\right)\right)(x, y)=
\end{gathered}
$$

$$
\begin{gathered}
=\frac{\partial\left(3 x^{2} y^{2}\right)}{\partial x}(x, y)=6 x y^{2} \\
D_{2} D_{1} f(x, y):=\frac{\partial}{\partial y}\left(\frac{\partial f}{\partial x}\right)(x, y):=\frac{\partial^{2} f}{\partial y \partial x}(x, y):=\left(D_{2}\left(D_{1} f\right)\right)(x, y)= \\
\frac{\partial\left(2 x y^{3}\right)}{\partial y}(x, y)=6 x y^{2} \\
D_{2} D_{2} f(x, y):=D_{2}^{2} f(x, y):=\frac{\partial^{2} f}{\partial y^{2}}(x, y):=\left(D_{2}\left(D_{2} f\right)\right)(x, y)= \\
=\frac{\partial\left(3 x^{2} y^{2}\right)}{\partial y}(x, y)=6 x^{2} y .
\end{gathered}
$$

Notice that

$$
2 y^{3}=\frac{\partial^{2} f}{\partial x^{2}}(x, y) \neq\left(\frac{\partial f}{\partial x}(x)\right)^{2}=\left(2 x y^{3}\right)^{2}=4 x^{2} y^{6}
$$

But we have that

$$
\frac{\partial}{\partial x}\left(\frac{\partial f}{\partial y}\right)(x, y)=6 x y^{2}=\frac{\partial}{\partial y}\left(\frac{\partial f}{\partial x}\right)(x, y) .
$$

This is not accident. One can show that if $U \subseteq \mathbb{R}^{2}$ is open and $f: U \rightarrow \mathbb{R}$ such that the partial derivatives

$$
D_{1} f(x, y), D_{2} f(x, y), D_{1} D_{2} f(x, y), D_{2} D_{1} f(x, y)
$$

exist and are continuous, then for every $(x, y) \in U$ we have that

$$
D_{1} D_{2} f(x, y)=D_{2} D_{1} f(x, y)
$$

We can have repeated partial derivatives for $n>2$. If $f: \mathbb{R}^{3} \rightarrow \mathbb{R}$ is defined by

$$
f(x, y, z)=x^{2} y z^{3}
$$

then

$$
D_{1} f(x, y, z)=2 x y z^{3} \quad D_{2} D_{1} f(x, y, z)=2 x z^{3} \quad D_{3} D_{2} D_{1} f(x, y, z)=6 x z^{2}
$$

and

$$
D_{3} f(x, y, z)=3 x^{2} y z^{2} \quad D_{2} D_{3} f(x, y, z)=3 x^{2} z^{2} \quad D_{1} D_{2} D_{3} f(x, y, z)=6 x z^{2}
$$

i.e.,

$$
D_{3} D_{2} D_{1} f(x, y, z)=6 x z^{2}=D_{1} D_{2} D_{3} f(x, y, z)
$$

By our previous remark on the equality $D_{1} D_{2} f(x, y)=D_{2} D_{1} f(x, y)$, if all the related partial derivatives exist and are continuous we get

$$
\begin{aligned}
D_{3} D_{2} D_{1} f(x, y, z) & =D_{3} D_{1} D_{2} f(x, y, z) \\
& =D_{1} D_{3} D_{2} f(x, y, z) \\
& =D_{1} D_{2} D_{3} f(x, y, z) .
\end{aligned}
$$

### 2.4. The chain rule

In this section we define when a function $f: U \rightarrow \mathbb{R}$, where $U$ is an open subset of $\mathbb{R}^{n}$, is differentiable at some point $x_{0} \in U$. To motivate this definition we notice the following fact.

Remark 2.4.1. Let $U$ be an open subset of $\mathbb{R}, x_{0} \in U$ and $f: U \rightarrow \mathbb{R}$. The following are equivalent:
(i) $f$ is differentiable at $x_{0}$.
(ii) There are $\epsilon>0, a \in \mathbb{R}$, and a function $g:(-\epsilon, \epsilon) \rightarrow \mathbb{R}$ such that

$$
f\left(x_{0}+h\right)-f\left(x_{0}\right)=a h+|h| g(h),
$$

for every $h \in(-\epsilon, \epsilon)$, and

$$
\lim _{h \rightarrow 0} g(h)=0
$$

Proof. If $f$ is differentiable at $x_{0}$, then

$$
a:=f^{\prime}\left(x_{0}\right):=\lim _{h \rightarrow 0} \frac{f\left(x_{0}+h\right)-f\left(x_{0}\right)}{h} \in \mathbb{R}
$$

and if $h \neq 0$, we define

$$
\phi(h)=\frac{f\left(x_{0}+h\right)-f\left(x_{0}\right)}{h}-f^{\prime}\left(x_{0}\right),
$$

while if $h=0$, we define $\phi(0):=0$. Clearly,

$$
\lim _{h \rightarrow 0} \phi(h)=0
$$

and for every $h$ in some $\epsilon$-interval around 0 we have that

$$
f\left(x_{0}+h\right)-f\left(x_{0}\right)=f^{\prime}\left(x_{0}\right) h+h \phi(h) .
$$

If we define $g(h):=\phi(h)$, if $h \geq 0$, and $g(h):=-\phi(h)$, if $h<0$, we have that

$$
|h| g(h)=h \phi(h)
$$

and we get the required equality

$$
f\left(x_{0}+h\right)-f\left(x_{0}\right)=a h+|h| g(h) .
$$

Of course,

$$
\lim _{h \rightarrow 0} g(h)=0
$$

For the converse, if $h \neq 0$, then

$$
\frac{f\left(x_{0}+h\right)-f\left(x_{0}\right)}{h}=\frac{a h+|h| g(h)}{h}=a+\frac{|h|}{h} g(h),
$$

which converges to $a$, as $h$ converges to 0 i.e., $a=f^{\prime}\left(x_{0}\right)$.

Definition 2.4.2. Let $U$ be an open subset of $\mathbb{R}^{n}, x_{0} \in U$ and $f: U \rightarrow \mathbb{R}$. We say that $f$ is differentiable at $x_{0}$, if
(a) The gradient of $f$ at $x_{0}$

$$
\operatorname{grad} f\left(x_{0}\right):=\left(D_{1} f\left(x_{0}\right), \ldots, D_{n} f\left(x_{0}\right)\right)=\left(\frac{\partial f}{\partial x_{1}}\left(x_{0}\right), \ldots, \frac{\partial f}{\partial x_{n}}\left(x_{0}\right)\right)
$$

exists, and
(b) there is a function $g$ defined on a small open ball around the origin $\mathbf{0}$ such that

$$
\lim _{|h| \rightarrow 0} g(h)=0
$$

and

$$
\begin{aligned}
f\left(x_{0}+h\right)-f\left(x_{0}\right) & =\frac{\partial f}{\partial x_{1}}\left(x_{0}\right) h_{1}+\ldots+\frac{\partial f}{\partial x_{n}}\left(x_{0}\right) h_{n}+|h| g(h) \\
& :=\left\langle(\operatorname{grad} f)\left(x_{0}\right), h\right\rangle+|h| g(h) .
\end{aligned}
$$

We say that $f$ is differentiable on $U$, if it is differentiable at every point of $U$.
To show that a function $f$ as above is differentiable on $U$, it suffices to show that the gradient of $f$ at every point of $U$ exists, and that the partial derivatives on $U$ are continuous functions (the proof is omitted).

Proposition 2.4.3. If $U$ is an open subset of $\mathbb{R}^{n}, x_{0} \in U$ and $f: U \rightarrow \mathbb{R}$, then $f$ is differentiable at $x_{0}$, if all partial derivatives of $f$ at $x_{0}$ exist in $U$ and for each $i \in\{1, \ldots, n\}$ the function

$$
U \ni x \mapsto \frac{\partial f}{\partial x_{i}}(x)
$$

is continuous at $x_{0}$.
Proof. See [4], p. 322.
In the one dimensional case the chain rule takes the form

$$
(f \circ g)^{\prime}(t)=f^{\prime}(g(t)) g^{\prime}(t)
$$

where $f$ and $g$ are as indicated in the following diagram


Next we prove the generalisation of this rule.
Proposition 2.4.4 (Chain rule). Let $I$ be an interval of $\mathbb{R}$, and $\boldsymbol{x}: I \rightarrow \mathbb{R}^{n}$ differentiable curve on $I$ such that $\boldsymbol{x}(I) \subseteq U$,

where $U$ is an open subset of $\mathbb{R}^{n}$. If $f: U \rightarrow \mathbb{R}$ is differentiable on $U$, then the function

$f \circ \boldsymbol{x}: I \rightarrow \mathbb{R}$ is differentiable, and for every $t \in I$ we have that

$$
(f \circ \boldsymbol{x})^{\prime}(t)=\left\langle(\operatorname{grad} f)(\boldsymbol{x}(t)), \boldsymbol{x}^{\prime}(t)\right\rangle .
$$

Proof. Let the quotient

$$
\frac{f(\boldsymbol{x}(t+h))-f(\boldsymbol{x}(t))}{h},
$$

which, if we define

$$
K:=K(t, h):=\boldsymbol{x}(t+h)-\boldsymbol{x}(t),
$$

and hence $\boldsymbol{x}(t+h)=K+\boldsymbol{x}(t)$, it becomes

$$
\frac{f(\boldsymbol{x}(t)+K)-f(\boldsymbol{x}(t))}{h}
$$

Since $f$ is differentiable on $U$, and $\boldsymbol{x}(t)$ is included in $U, f$ is differentiable at $\boldsymbol{x}(t)$, for every $t \in I$. By the Definition 2.4.2 there is a function $g$ such that

$$
f(\boldsymbol{x}(t)+K)-f(\boldsymbol{x}(t))=\langle(\operatorname{grad} f)(\boldsymbol{x}(t)), K\rangle+|K| g(K)
$$

and

$$
\lim _{|K| \rightarrow 0} g(K)=0
$$

Hence,

$$
\begin{aligned}
\frac{f(\boldsymbol{x}(t+h))-f(\boldsymbol{x}(t))}{h}= & \left\langle(\operatorname{grad} f)(\boldsymbol{x}(t)), \frac{\boldsymbol{x}(t+h)-\boldsymbol{x}(t)}{h}\right\rangle \\
& +\frac{|\boldsymbol{x}(t+h)-\boldsymbol{x}(t)|}{h} g(K)
\end{aligned}
$$

$$
\begin{aligned}
= & \left\langle(\operatorname{grad} f)(\boldsymbol{x}(t)), \frac{\boldsymbol{x}(t+h)-\boldsymbol{x}(t)}{h}\right\rangle \\
& \pm\left|\frac{\boldsymbol{x}(t+h)-\boldsymbol{x}(t)}{h}\right| g(K)
\end{aligned}
$$

If $h \rightarrow 0$, then

$$
\left\langle(\operatorname{grad} f)(\boldsymbol{x}(t)), \frac{\boldsymbol{x}(t+h)-\boldsymbol{x}(t)}{h}\right\rangle \rightarrow\left\langle(\operatorname{grad} f)(\boldsymbol{x}(t)), \boldsymbol{x}^{\prime}(t)\right\rangle,
$$

and

$$
\pm\left|\frac{\boldsymbol{x}(t+h)-\boldsymbol{x}(t)}{h}\right| g(K) \rightarrow \pm\left|\boldsymbol{x}^{\prime}(t)\right| 0=0
$$

since if $h \rightarrow 0$, then $K:=\boldsymbol{x}(t+h)-\boldsymbol{x}(t) \rightarrow 0$, and we use the fact that $\lim _{|K| \rightarrow 0} g(K)=0$.

Unfolding the chain rule we get

$$
\begin{aligned}
(f \circ \boldsymbol{x})^{\prime}(t) & =\left\langle\left(\frac{\partial f}{\partial x_{1}}(\boldsymbol{x}(t)), \ldots, \frac{\partial f}{\partial x_{n}}(\boldsymbol{x}(t))\right), \boldsymbol{x}^{\prime}(t)\right\rangle \\
& =\sum_{i=1}^{n} \frac{\partial f}{\partial x_{1}}(\boldsymbol{x}(t)) x_{i}{ }^{\prime}(t) \\
& =: \sum_{i} \frac{\partial f}{\partial x_{i}}(\boldsymbol{x}(t)) \frac{d x_{i}}{d t}(t) \\
& :=\frac{\partial f}{\partial x_{1}}(\boldsymbol{x}(t)) \frac{d x_{1}}{d t}(t)+\ldots+\frac{\partial f}{\partial x_{n}}(\boldsymbol{x}(t)) \frac{d x_{n}}{d t}(t),
\end{aligned}
$$

where $\boldsymbol{x}(t)=\left(x_{1}(t), \ldots, x_{n}(t)\right)$. For simplicity we also write

$$
\frac{d f(\boldsymbol{x}(t))}{d t}=\frac{\partial f}{\partial x_{1}} \frac{d x_{1}}{d t}+\ldots+\frac{\partial f}{\partial x_{n}} \frac{d x_{n}}{d t} .
$$

For example, let the following functions

defined by $\boldsymbol{x}(t):=\left(e^{t}, t, t^{2}\right)=(x(t), y(t), z(t))$ and $f(x, y, z):=x^{2} y z$. Then $f$ is differentiable on $\mathbb{R}^{3}$ by the Proposition 2.4.3, and by the chain rule we have that

$$
\begin{aligned}
\frac{d f(\boldsymbol{x}(t))}{d t} & =\frac{\partial f}{\partial x} \frac{d x}{d t}+\frac{\partial f}{\partial y} \frac{d y}{d t}+\frac{\partial f}{\partial z} \frac{d z}{d t} \\
& =2 x y z e^{t}+x^{2} z+x^{2} y 2 t .
\end{aligned}
$$

As a simple example of applying the chain rule, let $f: \mathbb{R}^{3} \rightarrow \mathbb{R}$ differentiable, and let $g: \mathbb{R} \rightarrow \mathbb{R}$, defined by

$$
g(t)=f(P+t Q)
$$

for every $t \in \mathbb{R}$, and some $P, Q \in \mathbb{R}^{3}$. In order to find $g^{\prime}(t)$, let $\boldsymbol{x}: \mathbb{R} \rightarrow \mathbb{R}^{3}$

such that $g=f \circ \boldsymbol{x}$. Let

$$
\boldsymbol{x}(t)=P+t Q=\left(p_{1}+t q_{1}, p_{2}+t q_{2}, p_{3}+t q_{3}\right)
$$

for every $t \in \mathbb{R}$. Since $\boldsymbol{x}^{\prime}(t)=\left(q_{1}, q_{2}, q_{3}\right)=Q$, we get

$$
g^{\prime}(t)=(f \circ \boldsymbol{x})^{\prime}(t)=\left\langle(\operatorname{grad} f)(\boldsymbol{x}(t)), \boldsymbol{x}^{\prime}(t)\right\rangle=\langle(\operatorname{grad} f)(P+t Q), Q\rangle
$$

Corollary 2.4.5. Let $U$ be an open subset of $\mathbb{R}^{n}$ such that for every two points $x_{0}, x_{1} \in U$ there is a differentiable curve $\boldsymbol{x}:[0,1] \rightarrow U$ such that $\boldsymbol{x}(0)=x_{0}$ and $\boldsymbol{x}(0)=x_{1}$.


If $f: U \rightarrow \mathbb{R}$ is differentiable on $U$, such that

$$
(\operatorname{grad} f)(x)=\mathbf{0},
$$

for every $x \in U$, then $f$ is constant on $U$.
Proof. Exercise.

### 2.5. Curve integrals

A vector field is a function $F: U \rightarrow \mathbb{R}^{n}$ that can be interpreted as a field of forces. If $\boldsymbol{x}: I \rightarrow U$ is a cure in $U$, the vector $\boldsymbol{x}(t)$ is interpreted as the position of the particle at time $t \in I$, and $F(\boldsymbol{x}(t))$ is the force acted upon the particle at position $\boldsymbol{x}(t)$. We may also say that the particle is moving in the force field $F$.


Definition 2.5.1. Let $U$ be an open subset of $\mathbb{R}^{n}$. A vector field on $U$ is a function $F: U \rightarrow \mathbb{R}^{n}$. If $F$ is represented by its coordinate functions i.e.,

$$
F=\left(f_{1}, \ldots, f_{n}\right),
$$

$F$ is differentiable on $U$, if each $f_{i}: U \rightarrow \mathbb{R}$ is differentiable on $U$. $F$ is called conservative, if there is a differentiable function $V: U \rightarrow \mathbb{R}$ such that ${ }^{1}$

$$
F=-\operatorname{grad} V
$$

In this case $V$ is called a potential energy function for $F$.
If $V$ is a potential energy function for $F$ and $c \in \mathbb{R}$ is some constant, then

$$
V+c
$$

is also a potential energy function for $F$. If $f$ is a differentiable function on $U$, then, by the Definition 2.4.2, we get the vector field on $U$ defined by

$$
U \ni x \mapsto(\operatorname{grad} f)(x)
$$

Let $F: U \rightarrow \mathbb{R}^{n}$ be a differentiable vector field on $U$ and $\boldsymbol{x}: I \rightarrow U$ a differentiable curve in $U$. Then the function $F \circ \boldsymbol{x}: I \rightarrow \mathbb{R}^{n}$ is well-defined

[^6]
and let the function on $I$ defined by
$$
t \mapsto\left\langle F(\boldsymbol{x}(t)), \boldsymbol{x}^{\prime}(t)\right\rangle,
$$
for every $t \in I$. E.g., let $F: \mathbb{R}^{2} \rightarrow \mathbb{R}^{2}$ defined by
$$
F(x, y):=\left(e^{x y}, y^{2}\right)
$$
for every $(x, y) \in \mathbb{R}^{2}$, and let $\boldsymbol{x}: \mathbb{R} \rightarrow \mathbb{R}^{2}$ be defined by
$$
\boldsymbol{x}(t):=(t, \sin t)
$$
for every $t \in \mathbb{R}$. Then
\[

$$
\begin{gathered}
\boldsymbol{x}^{\prime}(t)=(1, \cos t) \\
F(\boldsymbol{x}(t))=\left(e^{t \sin t}, \sin ^{2} t\right)
\end{gathered}
$$
\]

and

$$
\left\langle F(\boldsymbol{x}(t)), \boldsymbol{x}^{\prime}(t)\right\rangle=e^{t \sin t}+\left(\sin ^{2} t\right)(\cos t),
$$

for every $t \in \mathbb{R}$.
Definition 2.5.2. Let $U \subseteq \mathbb{R}^{n}$ be open, $\boldsymbol{x}:[a, b] \rightarrow U$ a differentiable curve with a differentiable derivative curve $\boldsymbol{x}^{\prime}$, and let $F: U \rightarrow \mathbb{R}^{n}$ be a differentiable vector field. The curve integral of $F$ along $\boldsymbol{x}$ is defined by

$$
\int_{\boldsymbol{x}} F:=\int_{a}^{b}\left\langle F(\boldsymbol{x}(t)), \boldsymbol{x}^{\prime}(t)\right\rangle d t .
$$

By the continuity of the inner product and our hypotheses on $\boldsymbol{x}$ and $F$ the function on $[a, b]$ defined by

$$
t \mapsto\left\langle F(\boldsymbol{x}(t)), \boldsymbol{x}^{\prime}(t)\right\rangle
$$

is continuous, hence Riemann-integrable. The above curve integral is a generalisation of the substitution method of the integral of functions in one variable:

$$
\int_{u(a)}^{u(b)} f(u) d u=\int_{a}^{b} f(u(t)) \frac{d u}{d t} d t .
$$

We use the following parametrisations of a linear, parabolic or circular segment: (I) If $P, Q \in \mathbb{R}^{n}$, the linear segment "from $P$ to $Q$ " is parametrised by the curve $\boldsymbol{x}:[0,1] \rightarrow \mathbb{R}^{n}$, defined by

$$
\boldsymbol{x}(t):=P+t(Q-P)
$$

for every $t \in[0,1]$. Clearly, $\boldsymbol{x}(0)=P$ and $\boldsymbol{x}(1)=Q$.
(II) A parabolic segment of the parabola $y=t^{2}$

is parametrised by the curve $\boldsymbol{x}(t):=\left(t, t^{2}\right)$, where $t$ is in a closed interval determined by the specifications of the respected problem.
(III) A parabolic segment of the parabola $x=t^{2}$

is parametrised by the curve $\boldsymbol{x}(t):=\left(t^{2}, t\right)$, where $t$ is in a closed interval determined by the specifications of the respected problem.
(IV) A circular segment of the circle of radius $r>0$ centered at $(0,0)$ in $\mathbb{R}^{2}$

is parametrised by the curve $\boldsymbol{x}(t):=(r \cos t, r \sin t)$, where $t$ is in a closed interval determined by the specifications of the respected problem.
Let the vector filed $F: \mathbb{R}^{2} \rightarrow \mathbb{R}^{2}$, defined by

$$
F(x, y):=\left(x^{2}, x y\right)
$$

for every $(x, y) \in \mathbb{R}^{2}$. To determine the integral of $F$ over the parabolic segment

from $P:=(-1,1)$ to $Q:=(1,1)$ we have that $\boldsymbol{x}(t)=\left(t, t^{2}\right)$ and $\boldsymbol{x}^{\prime}(t)=(1,2 t)$, and

$$
\begin{aligned}
& F(\boldsymbol{x}(t))=F\left(t, t^{2}\right)=\left(t^{2}, t^{3}\right), \\
& \left\langle F(\boldsymbol{x}(t)), \boldsymbol{x}^{\prime}(t)\right\rangle=t^{2}+2 t^{4},
\end{aligned}
$$

hence, since $-1 \leq t \leq 1$,

$$
\begin{aligned}
\int_{\boldsymbol{x}} F & =\int_{-1}^{1}\left(t^{2}+2 t^{4}\right) d t \\
& =\int_{-1}^{1} t^{2} d t+\int_{-1}^{1} 2 t^{4} d t \\
& =\frac{2}{3}+\frac{4}{5}
\end{aligned}
$$

To determine the curve integral of the vector field $F: \mathbb{R}^{2} \rightarrow \mathbb{R}^{2}$, defined by

$$
F(x, y):=\left(x^{2} y, y^{3}\right)
$$

for every $(x, y) \in \mathbb{R}^{2}$, over the line segment from $P:=(0,0)$ to $Q:=(1,1)$ we use the parametrisation of the segment

$$
\boldsymbol{x}(t)=P+t(Q-P)=(0,0)+t((1,1)-(0,0))=t(1,1)=(t, t),
$$

where $t \in[0,1]$, and hence $F(\boldsymbol{x}(t))=F(t, t)=\left(t^{3}, t^{3}\right), \boldsymbol{x}^{\prime}(t)=(1,1)$,

$$
\left\langle F(\boldsymbol{x}(t)), \boldsymbol{x}^{\prime}(t)\right\rangle=t^{3}+t^{3}=2 t^{3},
$$

and

$$
\int_{\boldsymbol{x}} F=\int_{0}^{1} 2 t^{3}=2 \int_{0}^{1} t^{3}=2 \frac{1}{4}=\frac{1}{2} .
$$

Let the vector field $F: \mathbb{R}^{2} \backslash\{(0,0)\} \rightarrow \mathbb{R}^{2}$, defined by

$$
F(x, y):=\left(\frac{-y}{x^{2}+y^{2}}, \frac{x}{x^{2}+y^{2}}\right)
$$

for every $(x, y)$ in the open subset $\mathbb{R}^{2} \backslash\{(0,0)\}$ of $\mathbb{R}^{2}$. To determine its integral over the circular segment of the circle of radius 3 around $(0,0)$ from $P:=(3,0)$ to

$$
Q:=\left(\frac{3 \sqrt{3}}{2}, \frac{3}{3}\right)
$$

we consider the curve

$$
\boldsymbol{x}(t)=(3 \cos t, 3 \sin t), \quad \boldsymbol{x}^{\prime}(t)=(-3 \sin t, 3 \cos t), \quad t \in\left[0, \frac{\pi}{6}\right],
$$

since $\boldsymbol{x}(0)=P$ and $\boldsymbol{x}\left(\frac{\pi}{6}\right)=Q$. Since

$$
\begin{aligned}
F(\boldsymbol{x}(t)) & =F(3 \cos t, 3 \sin t) \\
& =\left(\frac{-3 \sin t}{(3 \cos t)^{2}+(-3 \sin t)^{2}}, \frac{3 \cos t}{(3 \cos t)^{2}+(-3 \sin t)^{2}}\right) \\
& =\left(\frac{-3 \sin t}{9}, \frac{3 \cos t}{9}\right) \\
& =\frac{1}{3}(-\sin t, \cos t),
\end{aligned}
$$

and

$$
\left\langle F(\boldsymbol{x}(t)), \boldsymbol{x}^{\prime}(t)\right\rangle=\sin ^{2} t+\cos ^{2} t=1,
$$

we get

$$
\int_{\boldsymbol{x}} F=\int_{0}^{\frac{\pi}{6}} d t=\frac{\pi}{6}
$$

Definition 2.5.3. A path in an open subset $U$ of $\mathbb{R}^{n}$ is a finite sequence

$$
p:=\left(\boldsymbol{x}_{1}, \ldots, \boldsymbol{x}_{m}\right)
$$

where $m \geq 1, \boldsymbol{x}_{1}:\left[a_{1}, b_{1}\right] \rightarrow U, \ldots, \boldsymbol{x}_{m}:\left[a_{m}, b_{m}\right] \rightarrow U$ are curves in $U$ such that

$$
\boldsymbol{x}_{1}\left(b_{1}\right)=\boldsymbol{x}_{2}\left(a_{2}\right) \& \ldots \& \boldsymbol{x}_{m}\left(a_{m}\right)=\boldsymbol{x}_{m-1}\left(b_{m-1}\right) .
$$

A path $p$ is called differentiable on $U$, if $\boldsymbol{x}_{1}, \ldots, \boldsymbol{x}_{m}$ are differentiable curves on $U$ with differentiable derivative curves on $U$. We also say that $p$ is closed, if

$$
\boldsymbol{x}_{1}\left(a_{1}\right)=\boldsymbol{x}_{m}\left(b_{m}\right) .
$$

If $F: U \rightarrow \mathbb{R}^{n}$ is a differentiable vector field on $U$, and $p$ is a differentiable path on $U$, the path integral of $F$ over $p$ is defined by

$$
\int_{p} F:=\int_{\boldsymbol{x}_{1}} F+\ldots+\int_{\boldsymbol{x}_{m}} F
$$

Clearly, a curve in $U$ is a special case of a path in $U$.

### 2.6. Conservative vector fields

Theorem 2.6.1. Let $U \subseteq \mathbb{R}^{n}$ be open, and $F: U \rightarrow \mathbb{R}^{n}$ be a differentiable vector field on $U$.
(I) Let $F=\operatorname{grad} V$, for some differentiable function $V: U \rightarrow \mathbb{R}$.
(a) If $\boldsymbol{x}:[a, b] \rightarrow U$ is a differentiable curve in $U$ with $\boldsymbol{x}(a)=P$ and $\boldsymbol{x}(b)=Q$, then

$$
\int_{\boldsymbol{x}} F=V(Q)-V(P)
$$

(b) If $\boldsymbol{y}:[a, b] \rightarrow U$ is a differentiable curve in $U$ with $\boldsymbol{y}(a)=P$ and $\boldsymbol{y}(b)=Q$, then

$$
\int_{\boldsymbol{y}} F=\int_{\boldsymbol{x}} F
$$

(c) If $\boldsymbol{z}:[a, b] \rightarrow U$ is a closed differentiable curve in $U$ i.e., $\boldsymbol{z}(a)=P=\boldsymbol{z}(b)$, then

$$
\int_{z} F=0
$$

(II) Let $F=-\operatorname{grad} V$, for some differentiable function $V: U \rightarrow \mathbb{R}$.
(a) If $\boldsymbol{x}:[a, b] \rightarrow U$ is a differentiable curve in $U$ with $\boldsymbol{x}(a)=P$ and $\boldsymbol{x}(b)=Q$, then

$$
\int_{\boldsymbol{x}} F=V(P)-V(Q)
$$

(b) If $\boldsymbol{y}:[a, b] \rightarrow U$ is a differentiable curve in $U$ with $\boldsymbol{y}(a)=P$ and $\boldsymbol{y}(b)=Q$, then

$$
\int_{\boldsymbol{y}} F=\int_{\boldsymbol{x}} F
$$

(c) If $\boldsymbol{z}:[a, b] \rightarrow U$ is a closed differentiable curve in $U$ i.e., $\boldsymbol{z}(a)=P=\boldsymbol{z}(b)$, then

$$
\int_{z} F=0 .
$$

Proof. We prove only the first part of (i) and the rest is an exercise. By the definition of the curve integral of $f$ and the chain rule on $V \circ \boldsymbol{x}$

we have that

$$
\begin{aligned}
\int_{\boldsymbol{x}} F & :=\int_{a}^{b}\left\langle F(\boldsymbol{x}(t)), \boldsymbol{x}^{\prime}(t)\right\rangle d t \\
& =\int_{a}^{b}\left\langle(\operatorname{grad} V)(\boldsymbol{x}(t)), \boldsymbol{x}^{\prime}(t)\right\rangle d t \\
& =\int_{a}^{b}(V \circ \boldsymbol{x})^{\prime}(t) d t \\
& =[V \circ \boldsymbol{x}]_{a}^{b} \\
& =V(\boldsymbol{x}(b))-V(\boldsymbol{x}(a)) \\
& =V(Q)-V(P) .
\end{aligned}
$$

Because of the above independence of the integral $\int_{\boldsymbol{x}} F$ of a conservative vector field from the curve connecting the points $P$ and $Q$ in $U$, we write

$$
\int_{P}^{Q} F:=\int_{\boldsymbol{x}} F=V(Q)-V(P)
$$

where $\boldsymbol{x}$ is any curve in $U$ from $P$ to $Q$. Let $F: \mathbb{R}^{3} \rightarrow \mathbb{R}^{3}$ a vector field defined by

$$
F(x, y, z):=\left(2 x y^{3} z, 3 x^{2} y^{2} z, x^{2} y^{3}\right)
$$

for every $(x, y, z) \in \mathbb{R}^{3}$. If $V: \mathbb{R}^{3} \rightarrow \mathbb{R}$ is defined by

$$
V(x, y, z):=x^{2} y^{3} z
$$

it is easy to see that $F=\operatorname{grad} V$. If $P:=(1,-1,2)$ and $Q:=(-3,2,5)$, then

$$
\int_{P}^{Q} F=V(Q)-V(P)=V(-3,2,5)-V(1,-1,2)=360-(-2)=362
$$

Let $G: \mathbb{R}^{3} \backslash\{(0,0,0)\} \rightarrow \mathbb{R}^{3}$ be defined by

$$
G(x, y, z):=\frac{k(x, y, z)}{|(x, y, z)|^{3}}
$$

for every $(x, y, z) \in \mathbb{R}^{3} \backslash\{(0,0,0)\}$ and some $k \in \mathbb{R}$. If $V: \mathbb{R}^{3} \backslash\{(0,0,0)\} \rightarrow \mathbb{R}$ is defined by

$$
V(x, y, z):=-\frac{k}{|(x, y, z)|},
$$

for every $(x, y, z) \in \mathbb{R}^{3} \backslash\{(0,0,0)\}$, then one can show (exercise) that

$$
\operatorname{grad} V=G
$$

i.e.,

$$
\left(\frac{\partial V}{\partial x}(x, y, z), \frac{\partial V}{\partial y}(x, y, z), \frac{\partial V}{\partial z}(x, y, z)\right)=\frac{k}{|(x, y, z)|^{3}}(x, y, z) .
$$

If $P:=(1,1,1)$ and $Q:=(1,2,-1)$, then

$$
\begin{aligned}
\int_{P}^{Q} G & =V(Q)-V(P) \\
& =-\frac{k}{|Q|}-\left(-\frac{k}{|P|}\right) \\
& =-k\left(\frac{1}{|Q|}-\frac{1}{|P|}\right) \\
& =-k\left(\frac{1}{\sqrt{6}}-\frac{1}{\sqrt{3}}\right) .
\end{aligned}
$$

### 2.7. Green's theorem on rectangles

Definition 2.7.1. An rectangle $\mathcal{R}$ in $\mathbb{R}^{2}$ is a set of the form

$$
\mathcal{R}:=[a, b] \times[c, d]:=\left\{(x, y) \in \mathbb{R}^{2} \mid a \leq x \leq b \& c \leq y \leq d\right\}
$$

and an open rectangle is a set of the form

$$
\mathcal{R}^{o}:=(a, b) \times(c, d):=\left\{(x, y) \in \mathbb{R}^{2} \mid a<x<b \& c<y<d\right\}
$$

If $f: \mathcal{R} \rightarrow \mathbb{R}$ is a continuous function ${ }^{2}$, the double integral of $f$ on $\mathcal{R}$ is defined by

$$
\iint_{\mathcal{R}} f:=\int_{a}^{b}\left(\int_{c}^{d} f(x, y) d y\right) d x .
$$

[^7]Let $\mathcal{R}:=[1,2] \times[-3,4]$ and $f: \mathcal{R} \rightarrow \mathbb{R}$, defined by

$$
f(x, y):=x^{2} y
$$

for every $(x, y) \in \mathbb{R}^{2}$. Then

$$
\begin{aligned}
\iint_{\mathcal{R}} f & =\int_{1}^{2}\left(\int_{-3}^{4} x^{2} y d y\right) d x \\
& =\int_{1}^{2} x^{2}\left(\int_{-3}^{4} y d y\right) d x \\
& =\int_{1}^{2} x^{2} \frac{1}{2}(16-9) d x \\
& =\frac{7}{2} \int_{1}^{2} x^{2} d x \\
& =\frac{7}{2} \frac{1}{3}\left(2^{3}-1^{3}\right) \\
& =\frac{49}{6}
\end{aligned}
$$

If $U \subseteq \mathbb{R}^{2}$ is open, and let $F: U \rightarrow \mathbb{R}^{2}$ be a differentiable vector field on $U$ such that

$$
F(x, y):=(p(x, y), q(x, y)),
$$

where $p, q: U \rightarrow \mathbb{R}$ are the components of $F$. If $\boldsymbol{x}:[a, b] \rightarrow U$ is a differentiable curve in $U$, then

$$
\begin{aligned}
\int_{\boldsymbol{x}} F & :=\int_{a}^{b}\left\langle F(\boldsymbol{x}(t)), \boldsymbol{x}^{\prime}(t)\right\rangle d t \\
& =\int_{a}^{b}\left(p(x, y) \frac{d x}{d t}+q(x, y) \frac{d y}{d t}\right) d t \\
& =\int_{a}^{b} p(x, y) d x+q(x, y) d y
\end{aligned}
$$

According to the next theorem, if we want to find the path-integral

$$
\int_{\left(\boldsymbol{x}_{1}, \boldsymbol{x}_{2}, \boldsymbol{x}_{3}, \boldsymbol{x}_{4}\right)} F
$$

of a differentiable vector field $F=(p, q)$ defined on an open rectangle, where the path

$$
\left(\boldsymbol{x}_{1}, \boldsymbol{x}_{2}, \boldsymbol{x}_{3}, \boldsymbol{x}_{4}\right)
$$

parametrises counterclockwise the rectangle $\mathcal{R}$, it suffices to calculate the double integral

$$
\iint_{\mathcal{R}}\left(\frac{\partial q}{\partial x}-\frac{\partial p}{\partial y}\right)
$$

Hence we do not need to calculate the curve integrals separately i.e., to use the equality

$$
\int_{\left(\boldsymbol{x}_{1}, \boldsymbol{x}_{2}, \boldsymbol{x}_{3}, \boldsymbol{x}_{4}\right)} F=\int_{\boldsymbol{x}_{1}} F+\int_{\boldsymbol{x}_{2}} F+\int_{\boldsymbol{x}_{3}} F+\int_{\boldsymbol{x}_{4}} F .
$$

If e.g., we consider the rectangle $[-1,1] \times[-1,1]$

a path that parametrises it counterclockwise is the following sequence of linear segments

$$
((-1,-1) \rightarrow(1,-1),(1,-1) \rightarrow(1,1),(1,1) \rightarrow(-1,1),(-1,1) \rightarrow(-1,-1))
$$

The proof of the next theorem is omitted.
Theorem 2.7.2 (Green's theorem on rectangles). Let $F:(a, b) \times(c, d) \rightarrow \mathbb{R}^{2}$ be a differentiable vector field on the open rectangle $(a, b) \times(c, d)$ such that

$$
F(x, y):=(p(x, y), q(x, y)),
$$

for every $(x, y) \in(a, b) \times(c, d)$. Then

$$
\int_{\left(\boldsymbol{x}_{1}, \boldsymbol{x}_{2}, \boldsymbol{x}_{3}, \boldsymbol{x}_{4}\right)} p(x, y) d x+q(x, y) d y=\iint_{\mathcal{R}}\left(\frac{\partial q}{\partial x}-\frac{\partial p}{\partial y}\right) .
$$

Let the vector field $F: \mathbb{R}^{2} \rightarrow \mathbb{R}^{2}$, defined by

$$
F(x, y)=\left(3 x y, x^{2}\right)
$$

for every $(x, y) \in \mathbb{R}^{2}$. Hence,

$$
\begin{aligned}
p(x, y) & =3 x y, \quad q(x, y)=x^{2} \\
\frac{\partial q}{\partial x} & =2 x \quad \& \quad \frac{\partial p}{\partial y}=3 x
\end{aligned}
$$

The integral of $F$ around the following rectangle

is calculated with the use of Green's theorem as follows

$$
\begin{aligned}
\int_{\left(\boldsymbol{x}_{1}, \boldsymbol{x}_{2}, \boldsymbol{x}_{3}, \boldsymbol{x}_{4}\right)} p(x, y) d x+q(x, y) d y & =\iint_{\mathcal{R}}\left(\frac{\partial q}{\partial x}-\frac{\partial p}{\partial y}\right) \\
& =\int_{-1}^{3}\left(\int_{0}^{2}(2 x-3 x) d y\right) d x \\
& =\int_{-1}^{3}\left(\int_{0}^{2}(-x) d y\right) d x \\
& =\int_{-1}^{3}(-x)\left(\int_{0}^{2} d y\right) d x \\
& =\int_{-1}^{3}(-x) 2 d x \\
& =-2 \int_{-1}^{3} x d x \\
& =-8
\end{aligned}
$$

## CHAPTER 3

## Appendix

### 3.1. Solution to Exercise 2(ii), Sheet 1

If $f: \mathbb{R} \rightarrow \mathbb{R}$, we say that $f$ is differentiable at $x_{0} \in \mathbb{R}$, if there is some $l \in \mathbb{R}$ such that

$$
l=\lim _{h \rightarrow 0} \frac{f\left(x_{0}+h\right)-f\left(x_{0}\right)}{h},
$$

where using the $(\epsilon-\delta)$-definition of the notion of limit, this means that

$$
\forall_{\epsilon>0} \exists_{\delta_{f}(\epsilon)>0} \forall_{h \neq 0}\left(|h|<\delta_{f}(\epsilon) \Rightarrow\left|\frac{f\left(x_{0}+h\right)-f\left(x_{0}\right)}{h}-l\right| \leq \epsilon\right) .
$$

This necessarily unique limit $l$ is called the derivative of $f$ at $x_{0}$, and it is denoted by $f^{\prime}\left(x_{0}\right)$. The function $f$ is called continuous at $x_{0}$, if

$$
\lim _{h \rightarrow 0} f\left(x_{0}+h\right)=f\left(x_{0}\right) \Leftrightarrow \lim _{h \rightarrow 0}\left[f\left(x_{0}+h\right)-f\left(x_{0}\right)\right]=0
$$

where using the $(\epsilon-\delta)$-definition of the notion of limit, this means that

$$
\forall_{\epsilon>0} \exists_{\delta_{f}(\epsilon)>0} \forall_{h \in \mathbb{R}}\left(|h|<\delta_{f}(\epsilon) \Rightarrow\left|f\left(x_{0}+h\right)-f\left(x_{0}\right)\right| \leq \epsilon\right)
$$

If $f$ is differentiable at $x_{0}$, then $f$ is continuous at $x_{0}$. To show this we remark that the function

$$
\phi(h):= \begin{cases}\frac{f\left(x_{0}+h\right)-f\left(x_{0}\right)}{h} & , h \neq 0 \\ f^{\prime}\left(x_{0}\right)^{h} & , h=0\end{cases}
$$

is continuous at 0 . Since for $h \neq 0$ we have that

$$
f\left(x_{0}+h\right)-f\left(x_{0}\right)=h \phi(h) \Rightarrow f\left(x_{0}+h\right)=h \phi(h)+f\left(x_{0}\right),
$$

we get

$$
\lim _{h \rightarrow 0} f\left(x_{0}+h\right)=\lim _{h \rightarrow 0}\left[h \phi(h)+f\left(x_{0}\right)\right]=0 f^{\prime}\left(x_{0}\right)+f\left(x_{0}\right)=f\left(x_{0}\right) .
$$

The function $f$ is called differentiable, if it is differentiable at every $x_{0} \in \mathbb{R}$, and it is called continuous, if it is continuous at every $x_{0} \in \mathbb{R}$. If

$$
\begin{gathered}
C(\mathbb{R}):=\{f: \mathbb{R} \rightarrow \mathbb{R} \mid f \text { is continuous }\} \\
D(\mathbb{R}):=\{f: \mathbb{R} \rightarrow \mathbb{R} \mid f \text { is differentiable }\}
\end{gathered}
$$

the previous remark implies that

$$
D(\mathbb{R}) \subseteq C(\mathbb{R})
$$

Next we show that $D(\mathbb{R})$ is a linear space, and for $C(\mathbb{R})$ we work similarly. Clearly, the constant function 0 is differentiable and its derivative is at every $x_{0} \in \mathbb{R}$ again 0 . Next we show that if $f, g \in D(\mathbb{R})$, then $f+g \in D(\mathbb{R})$. Let $x_{0} \in \mathbb{R}$. Suppose that

$$
l=\lim _{h \rightarrow 0} \frac{f\left(x_{0}+h\right)-f\left(x_{0}\right)}{h} \quad \& \quad \lim _{h \rightarrow 0} \frac{g\left(x_{0}+h\right)-g\left(x_{0}\right)}{h}=m
$$

Using the triangle inequality

$$
|a+b| \leq|a|+|b|
$$

where $a, b \in \mathbb{R}$, we have that

$$
\begin{gathered}
\left|\frac{(f+g)\left(x_{0}+h\right)-(f+g)\left(x_{0}\right)}{h}-(l+m)\right|= \\
\left|\frac{f\left(x_{0}+h\right)+g\left(x_{0}+h\right)-f\left(x_{0}\right)-g\left(x_{0}\right)}{h}-l-m\right|= \\
\left|\frac{f\left(x_{0}+h\right)-f\left(x_{0}\right)}{h}-l+\frac{g\left(x_{0}+h\right)-g\left(x_{0}\right)}{h}-m\right| \leq \\
\left|\frac{f\left(x_{0}+h\right)-f\left(x_{0}\right)}{h}-l\right|+\left|\frac{g\left(x_{0}+h\right)-g\left(x_{0}\right)}{h}-m\right|,
\end{gathered}
$$

and since these two terms become arbitrarily small, for appropriate $h$, we get that

$$
(f+g)^{\prime}\left(x_{0}\right)=l+m=f^{\prime}\left(x_{0}\right)+g^{\prime}\left(x_{0}\right),
$$

and since $x_{0}$ is an arbitrary real number, we conclude that $f+g \in D(\mathbb{R})$. Finally, we show that if $a \in \mathbb{R}$ and $f \in D(\mathbb{R})$, then $a \cdot f \in D(\mathbb{R})$. Since $|a b|=|a||b|$, for every $a, b \in \mathbb{R}$, we have that

$$
\begin{gathered}
\left|\frac{(a \cdot f)\left(x_{0}+h\right)-(a \cdot f)\left(x_{0}\right)}{h}-a l\right|=\left|a\left(\frac{f\left(x_{0}+h\right)-f\left(x_{0}\right)}{h}-l\right)\right| \\
=|a|\left|\frac{f\left(x_{0}+h\right)-f\left(x_{0}\right)}{h}-l\right|
\end{gathered}
$$

and since the right term becomes arbitrarily small for appropriate $h$, we get

$$
(a \cdot f)^{\prime}\left(x_{0}\right)=a l=a f^{\prime}\left(x_{0}\right)
$$

Since $x_{0} \in \mathbb{R}$ is arbitrary, we conclude that $a \cdot f \in D(\mathbb{R})$.
3.2. On the solution of the Exercise $4(i)$, Sheet 3

The fact that $\int f \in D(\mathbb{R})$ is explained by the following fundamental result.
Theorem 3.2.1. Let $a, b, c, d \in \mathbb{R}$ such that $a \leq b \leq c \leq d$, and $f:[a, d] \rightarrow \mathbb{R}$ continuous. The function $\phi:[a, d] \rightarrow \mathbb{R}$, defined by

$$
\phi(x):=\int_{a}^{x} f(t) d t
$$

for every $x \in[a, d]$ is differentiable in $[a, d]$ and $\phi^{\prime}(x)=f(x)$.
Proof. We will use the following two basic properties of the Riemann integral. (1) If $m \leq f(t) \leq M$, for every $t \in[b, c]$, then

$$
m(c-b) \leq \int_{b}^{c} f(t) d t \leq M(c-b)
$$

$$
\begin{equation*}
\int_{a}^{c} f(t) d t=\int_{a}^{b} f(t) d t+\int_{b}^{c} f(t) d t \tag{2}
\end{equation*}
$$

If $x_{0} \in[a, d]$, then by (2) we have that

$$
\begin{aligned}
\frac{\phi\left(x_{0}+h\right)-\phi\left(x_{0}\right)}{h} & :=\frac{\int_{a}^{x_{0}+h} f(t) d t-\int_{a}^{x_{0}} f(t) d t}{h} \\
& =\frac{\int_{a}^{x_{0}} f(t) d t+\int_{x_{0}}^{x_{0}+h} f(t) d t-\int_{a}^{x_{0}} f(t) d t}{h} \\
& =\frac{\int_{x_{0}}^{x_{0}+h} f(t) d t}{h} .
\end{aligned}
$$

Since $f$ is continuous on the compact interval $\left[x_{0}, x_{0}+h\right]$, let $s, s^{\prime} \in\left[x_{0}, x_{0}+h\right]$ such that

$$
\begin{aligned}
f(s) & :=\min \left\{f(t) \mid t \in\left[x_{0}, x_{0}+h\right]\right\}:=m \\
f\left(s^{\prime}\right) & :=\max \left\{f(t) \mid t \in\left[x_{0}, x_{0}+h\right]\right\}:=M
\end{aligned}
$$

By (1) we have that

$$
\begin{aligned}
& m\left(x_{0}+h-x_{0}\right) \leq \int_{x_{0}}^{x_{0}+h} f(t) d t \leq M\left(x_{0}+h-x_{0}\right) \Leftrightarrow \\
& f(s) h \leq \int_{x_{0}}^{x_{0}+h} f(t) d t \leq f\left(s^{\prime}\right) h \stackrel{h \neq 0}{\Rightarrow} \\
& f(s) \leq \frac{\int_{x_{0}}^{x_{0}+h} f(t) d t}{h} \leq f\left(s^{\prime}\right)
\end{aligned}
$$

If $h \rightarrow 0$, then $s, s^{\prime} \rightarrow x_{0}$, and by the continuity of $f$ we get $f(s) \rightarrow f\left(x_{0}\right)$ and $f\left(s^{\prime}\right) \rightarrow f\left(x_{0}\right)$. By the sandwich lemma we get

$$
\phi^{\prime}\left(x_{0}\right):=\lim _{h \rightarrow 0} \frac{\phi\left(x_{0}+h\right)-\phi\left(x_{0}\right)}{h}=\lim _{h \rightarrow 0} \frac{\int_{x_{0}}^{x_{0}+h} f(t) d t}{h}=f\left(x_{0}\right) .
$$

### 3.3. Solution to Exercise 4(iv)-(v), Sheet 4

Let $A:=\left[a_{i j}\right] \in M_{m, n}(\mathbb{R}), B:=\left[b_{j k}\right] \in M_{n, l}(\mathbb{R})$ and $D:=\left[d_{k r}\right] \in M_{l, s}(\mathbb{R})$. (iv) By the definition of the multiplication of matrices we have that

$$
\begin{gathered}
A B:=\left[a_{i j}\right]\left[b_{j k}\right]:=\left[\sum_{j=1}^{n} a_{i j} b_{j k}\right], \\
B D:=\left[b_{j k}\right]\left[d_{k r}\right]:=\left[\sum_{k=1}^{l} b_{j k} d_{k r}\right], \\
(A B) D:=\left[\sum_{j=1}^{n} a_{i j} b_{j k}\right]\left[d_{k r}\right]:=\left[\sum_{k=1}^{l}\left(\sum_{j=1}^{n} a_{i j} b_{j k}\right) d_{k r}\right], \\
A(B D):=\left[a_{i j}\right]\left[\sum_{k=1}^{l} b_{j k} d_{k r}\right]:=\left[\sum_{j=1}^{n} a_{i j}\left(\sum_{k=1}^{l} b_{j k} d_{k r}\right)\right] .
\end{gathered}
$$

Since

$$
\sum_{k=1}^{l}\left(\sum_{j=1}^{n} a_{i j} b_{j k}\right) d_{k r}=\sum_{j=1}^{n} a_{i j}\left(\sum_{k=1}^{l} b_{j k} d_{k r}\right)=\sum_{k=1}^{l} \sum_{j=1}^{n} a_{i j} b_{j k} d_{k r}
$$

we get that $(A B) D=A(B D)$.
(v) Since

$$
\begin{aligned}
& A^{t}:=\left[\alpha_{j i}\right] \in M_{n, m}(\mathbb{R}), \quad \alpha_{j i}:=a_{i j}, \\
& B^{t}:=\left[\beta_{k j}\right] \in M_{l, n}(\mathbb{R}), \quad \beta_{k j}:=b_{j k},
\end{aligned}
$$

the product $B^{t} A^{t} \in M_{l, m}(\mathbb{R})$ is well-defined. Moreover, we have that

$$
(A B)^{t}:=\left[\sum_{j=1}^{n} a_{i j} b_{j k}\right]^{t}:=\left[\gamma_{k i}\right]
$$

where

$$
\gamma_{k i}:=\sum_{j=1}^{n} a_{i j} b_{j k}
$$

Since

$$
B^{t} A^{t}:=\left[\beta_{k j}\right]\left[\alpha_{j i}\right]:=\left[\sum_{j=1}^{n} \beta_{k j} \alpha_{j i}\right]=\left[\sum_{j=1}^{n} b_{j k} a_{i j}\right]=\left[\sum_{j=1}^{n} a_{i j} b_{j k}\right]
$$

we get $B^{t} A^{t}=(A B)^{t}$.

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[^0]:    ${ }^{1}$ A field is a structure $(\mathbb{F} ;+, \mathbf{0}, \cdot, \mathbf{1})$, where $\mathbb{F}$ is a set, $\mathbf{0}, \mathbf{1} \in \mathbb{F},+: \mathbb{F} \times \mathbb{F} \rightarrow \mathbb{F}$, and $\cdot: \mathbb{F} \times \mathbb{F} \rightarrow \mathbb{F}$ such that together with $\left(\mathrm{LS}_{1}\right)-\left(\mathrm{LS}_{4}\right)$ the following conditions are satisfied:
    $\forall_{x, y, z \in \mathbb{F}}(x \cdot(y \cdot z)=(x \cdot y) \cdot z)$.
    $\forall_{x, y, z \in \mathbb{F}}(x \cdot(y+z)=x \cdot y+x \cdot z)$.
    $\forall_{x, y \in \mathbb{F}}(x \cdot y=y \cdot x)$.
    $\forall_{x \in \mathbb{F}}(\mathbf{1} \cdot x=x)$.

[^1]:    ${ }^{2}$ This is the logical principle $P \vee \neg P$, where $P$ is any well-formed formula.

[^2]:    ${ }^{3}$ The restriction $g_{\mid Y}$ of $g$ is the function $g_{\mid Y}: Y \rightarrow Z$, where $g_{\mid Y}(y):=g(y)$, for every $y \in Y$. Clearly, if $Y$ is a subspace of a linear space $X$ and $f \in \mathcal{L}(X, Z)$, then $f_{Y} \in \mathcal{L}(Y, Z)$.

[^3]:    ${ }^{4}$ The symbol $\delta_{k i}$ is known as Kronecker's delta.

[^4]:    ${ }^{5}$ A simpler argument is the following. As we have shown after the Definition 1.5.1, $T_{A}\left(e_{i}\right)$ is the $i$-column of $A$. Hence, $T_{A_{T}}\left(e_{i}\right)$ is the $i$-column of $A_{T}$, which is exactly $T\left(e_{i}\right)$ by the definition of $A_{T}$.

[^5]:    ${ }^{6}$ The following figure also explains why $|x+y| \leq|x|+|y|$.

[^6]:    ${ }^{1}$ The negative sign is only traditional, and it can be avoided.

[^7]:    ${ }^{2}$ All functions defined on a rectangle that we are going to study here are going to be continuous.

