# Chapter 0. Introduction to the Mathematical Method 

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## Outline

(0) Mathematical language

- Axioms, postulates, definitions and propositions (a)
- Logical operators (a)
- Qualifiers (b)
- Mathematical proofs
- Modus ponens (b)
- Modus tollens (c)
- Reductio ad absurdum (c)
- Induction (c)
- Case distinction (c)
- Counterexample (c)
- Common math mistakes (c)


## References


M. de Guzmán Ozámiz. Cómo hablar, demostrar y resolver en Matemáticas. Anaya (2003)

## A little bit of history

Modern logic is based on precise calculus rules and was born in the middle of the XIX ${ }^{\text {th }}$ century with Gottfried Leibniz (1847), George Boole (1847), Augustus de Morgan (1847) and Bertrand Russell (1910).


To know more about the history of logic visit

- http://individual.utoronto.ca/pking/miscellaneous/ history-of-logic.pdf
- http://en.wikipedia.org/wiki/History_of_logic


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## Axioms, postulates and propositions

## Axioms, postulates and propositions

Mathematical language has to be uniform (everybody must use it in the same way) and univocal (i.e., without any kind of ambiguity). We start from some initial statements called axioms, postulates and definitions. These elements are not questioned, they are not true or false, they simply are, and they serve to build a logical reasoning.

## Example

Axiom If $A$ and $B$ are equal to $C$, then $A$ is equal to $B$.
Postulate For any two points, there is a unique straight line that joins them.
Definition A prime number is a natural number that can only be divided by 1 and itself.

## Propositions

## Propositions

Based on axioms, postulates and definitions, we can construct propositions that are statements that refer to already introduced objects. Propositions can be true or false. They are named with capital letters A, B, C, ...

## Example

$$
\begin{aligned}
& \quad 2+3 \text { (is not a proposition) } \\
& \text { A: } 2+3=5 \text { (is a true proposition) } \\
& \text { B: } 2+3=7 \text { (is a false proposition) }
\end{aligned}
$$

## Construction of new propositions

We can construct new propositions using already existing ones and logical operators

## Example

$$
\begin{aligned}
& \text { A: } 2+2=4 \text { (true) } \\
& \text { B: } 2+3=5 \text { (true) } \\
& \text { C: } 2+3=7 \text { (false) } \\
& \text { D: A y B (true) } \\
& \text { E: A o C (true) }
\end{aligned}
$$

and quantifiers

## Example

A: Some numbers are prime (true)
B: All even numbers can be divided by 2 (true)
C: None of odd numbers can be divided by 2 (true)

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## $\bar{A}(\operatorname{not} A)$

## Definition

$\bar{A}$ is true if $A$ is false, and $\bar{A}$ is false if $A$ is true.

## Example

$$
\begin{aligned}
& \text { A: } 3+2=5 \text { (true) } \\
& \text { B: } \neg A \equiv 3+2 \neq 5 \text { (false) } \\
& \text { C: } 3+2=6 \text { (false) } \\
& \text { D: } \neg C \equiv 3+2 \neq 6 \text { (true) }
\end{aligned}
$$

Truth table

| $A$ | $\bar{A}$ |
| :---: | :---: |
| F | T |
| T | F |

## Properties

$$
\overline{\bar{A}}=A
$$

## $\bar{A}(\operatorname{not} A)$

A double negation is a positive statement.

## Example

$$
\begin{aligned}
& \text { A: } 3+2=5 \text { (true) } \\
& \text { B: } \bar{A} \equiv 3+2 \neq 5 \text { (false) } \\
& \text { C: } \bar{B} \equiv 3+2=5 \text { (true) }
\end{aligned}
$$

## Example

It is not true that John is not at home.
A: John is at home
B: $\bar{A} \equiv \operatorname{Not}(J o h n$ is at home) $\equiv$ John is not at home
C: $\bar{B} \equiv \operatorname{Not}(J o h n$ is not at home) $\equiv$ John is at home $\equiv \mathrm{A}$
If C is true, then A is true. Therefore, John is at home.

## $A \cap B(A$ and $B)$

Truth table

| $A$ | $B$ | $A \cap B$ |
| :---: | :---: | :---: |
| F | F | F |
| F | T | F |
| T | F | F |
| T | T | T |

## Properties

 $A \cap B=B \cap A$
## Example

$$
\begin{aligned}
& \text { A: } 3+2=5 \text { (true) } \\
& \text { B: } 2+2=4 \text { (true) } \\
& \text { C: } A \cap B \equiv 3+2=5 \text { and } 2+2=4 \text { (true) } \\
& \text { D: } 3+2=6 \text { (false) } \\
& \text { E: } D \cap B \equiv 3+2=6 \text { and } 2+2=4 \text { (false) }
\end{aligned}
$$

## $A \cap B(A$ and $B)$

The common language AND is sometimes equivalent to the mathematical AND

## Example

Triangle $A B C$ and triangle $A^{\prime} B^{\prime} C^{\prime}$ are equilateral $\Rightarrow$
$A$ : $A B C$ is equilateral
$B$ : $A^{\prime} B^{\prime} C^{\prime}$ is equilateral
$C$ : $A \cap B \equiv$ Triangle $A B C$ is equilateral AND Triangle $A^{\prime} B^{\prime} C^{\prime}$ is equilateral
and sometimes not

## Example

Triangle $A B C$ and triangle $A^{\prime} B^{\prime} C^{\prime}$ are similar $\nRightarrow$
$A$ : $A B C$ is similar
$B$ : $A^{\prime} B^{\prime} C^{\prime}$ is similar
C: $A \cap B \equiv$ Triangle $A B C$ is similar AND Triangle $A^{\prime} B^{\prime} C^{\prime}$ is similar

## $A \cup B(A$ or $B ; A$ and $/$ or $B)$

$15 \%$ discounts for customers having a student card or university card. Of course, people with both cards have a $15 \%$ discount. Inclusive OR.

## Truth table

| $A$ | $B$ | $A \cup B$ |
| :---: | :---: | :---: |
| F | F | F |
| F | T | T |
| T | F | T |
| T | T | T |

## Properties

$A \cup B=B \cup A$

## Example

$$
\begin{aligned}
& \text { A: } 3+2=5 \text { (true) } \\
& \text { B: } 2+2=4 \text { (true) } \\
& \text { C: } A \cup B \equiv 3+2=5 \text { or } 2+2=4 \text { (true) } \\
& \text { D: } 3+2=6 \text { (false) } \\
& \text { E: } D \cup B \equiv 3+2=6 \text { or } 2+2=4 \text { (true) }
\end{aligned}
$$

## $A \oplus B$ (either $A$ or $B ; A$ xor $B$ (eXclusive or))

We'll go to Paris or Berlin. Either Paris or Berlin, we cannot go to both places at the same time. Exclusive OR.

Truth table

| $A$ | $B$ | $A \oplus B$ |
| :---: | :---: | :---: |
| F | F | F |
| F | T | T |
| T | F | T |
| T | T | F |

## Properties

$$
A \oplus B=B \oplus A
$$

## Example

$$
\begin{aligned}
& \text { A: } a<5 \\
& \text { B: } a=5 \\
& \text { C: } A \oplus B \equiv a \leq 5
\end{aligned}
$$

If $a=3$, then $C$ is true. If $a=6$, then $C$ is false.

## Negation of and

## $\overline{A \cap B}=\bar{A} \cup \bar{B}$

This is one of Morgan's laws.

| $A$ | $B$ | $A \cap B$ | $\overline{A \cap B}$ | $\bar{A}$ | $\bar{B}$ | $\bar{A} \cup \bar{B}$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| F | F | F | $\mathbf{T}$ | T | T | $\mathbf{T}$ |
| F | T | F | $\mathbf{T}$ | T | F | $\mathbf{T}$ |
| T | F | F | $\mathbf{T}$ | F | T | $\mathbf{T}$ |
| T | T | T | $\mathbf{F}$ | F | F | $\mathbf{F}$ |

## Example

A: It rained on Monday
B: It rained on Tuesday
C: $\overline{A \cap B} \equiv \mathrm{It}$ is not true that it rained on both days $\equiv$ Either it did not rain on Monday or it did not rain on Tuesday.

## Negation of or

Inclusive OR: $\overline{A \cup B}=\bar{A} \cap \bar{B}$
This is another Morgan's law.

| $A$ | $B$ | $A \cup B$ | $\overline{A \cup B}$ | $\bar{A}$ | $\bar{B}$ | $\bar{A} \cap \bar{B}$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| F | F | F | $\mathbf{T}$ | T | T | $\mathbf{T}$ |
| F | T | T | $\mathbf{F}$ | T | F | $\mathbf{F}$ |
| T | F | T | $\mathbf{F}$ | F | T | $\mathbf{F}$ |
| T | T | T | $\mathbf{F}$ | F | F | $\mathbf{F}$ |

Exclusive OR: $\overline{A \oplus B}=(\bar{A} \cap B) \cup(A \cap \bar{B})$

| $A$ | $B$ | $A \oplus B$ | $\overline{A \oplus B}$ | $\bar{A}$ | $\bar{B}$ | $\bar{A} \cap B$ | $A \cap \bar{B}$ | $(\bar{A} \cap B) \cup(A \cap \bar{B})$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| F | F | F | T | T | T | F | F | $\mathbf{T}$ |
| F | T | T | F | T | F | T | F | $\mathbf{F}$ |
| T | F | T | F | F | T | F | T | $\mathbf{F}$ |
| T | T | F | T | F | F | F | F | $\mathbf{T}$ |

## $A \Rightarrow B(A$ implies $B)$

## Natural language

- A implies B
- $A$ is sufficient for $B$
- A guarantees B
- $B$ is necessary for $A$
- If $A$, then $B$
- If not $B$, then not $A$
Truth table

| $A$ | $B$ | $A \Rightarrow B$ |
| :---: | :---: | :---: |
| F | F | T |
| F | T | T |
| T | F | F |
| T | T | T |

## $A \Rightarrow B(A$ implies $B)$

In natural language "If ..., then ..." is not used in the mathematical sense.

## Example

If it rains, I'll stay at home.

If he is at home, is it raining?
We don't know, he didn't say what he would do if it was not raining.

## Example

I'm going to the bank. If it is open, I'll bring 1000 euros.

If he is back with 1000 euros, is the bank open?
We don't know, maybe a very good friend of his gave him 1000 euros.

## $A \Rightarrow B($ If not $B$, then not $A)$

## Example

I'm going to the bank. If it is open, I'll bring 1000 euros.
If I'm back without 1000 euros, is the bank open?
No, let's see why
A: Bank is open
B: I bring 1000 euros

| $A$ | $B$ | $A \Rightarrow B$ | Why |
| :---: | :---: | :---: | :--- |
| F | F | T | The bank was closed |
| F | T | T | A friend gave me |
| T | F | F | I lied |
| T | T | T | I withdrew 1000 euros from bank |

There is only one situation in which my statement is true (I did not lie) and in which I do not bring 1000 euros ( B is false) that is when the bank is closed ( A is also false).

## $A \Rightarrow B($ If not $B$, then not $A)$

We can generally formulate this analysis as

## Properties

$$
A \Rightarrow B=\bar{B} \Rightarrow \bar{A}
$$

| $A$ | $B$ | $A \Rightarrow B$ | $\bar{B}$ | $\bar{A}$ | $\bar{B} \Rightarrow \bar{A}$ |
| :---: | :---: | :---: | :---: | :---: | :---: |
| F | F | $\mathbf{T}$ | T | T | $\mathbf{T}$ |
| F | T | $\mathbf{T}$ | F | T | $\mathbf{T}$ |
| T | F | $\mathbf{F}$ | T | F | $\mathbf{F}$ |
| T | T | $\mathbf{T}$ | F | F | $\mathbf{T}$ |

## $A \Rightarrow B(\operatorname{Not}(A$ and not $B))$

Another interesting property

## Properties

$$
\begin{aligned}
& A \Rightarrow B=\overline{A \cap \bar{B}} \\
& \overline{A \Rightarrow B}=A \cap \bar{B}
\end{aligned}
$$

The proof of these properties is left to the reader.

## Example

I'm going to the bank. If it is open, I'll bring 1000 euros.
It is equivalent to:

It will not be the case that (the bank is open (A) and I don't bring 1000 euros (not B)).

## $A \Leftrightarrow B$ (A if and only if $B$ )

Truth table

$$
\begin{array}{cc|c}
A & B & A \Leftrightarrow B \\
\hline \mathrm{~F} & \mathrm{~F} & \mathrm{~T} \\
\mathrm{~F} & \mathrm{~T} & \mathrm{~F} \\
\mathrm{~T} & \mathrm{~F} & \mathrm{~F} \\
\mathrm{~T} & \mathrm{~T} & \mathrm{~T}
\end{array}
$$

## Properties

$$
A \Leftrightarrow B=(A \Rightarrow B) \cap(B \Rightarrow A)
$$

In plain language, we say:
A is necessary and sufficient for B
$B$ is necessary and sufficient for $A$

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## Qualifiers

## Example

There might be a person that reads all newspapers every day.
Every day, there might be a person that reads all newspapers.
Every one reads a newspaper every day.
Every day, there is a newspaper that everybody reads.

## Example

We say that the limit of the function $f(x)$ when $x$ goes to $x_{0}$ is $y$ if and only if for all positive numbers $(\epsilon)$, there exists another positive number $(\delta)$ such that if the distance between $x$ and $x_{0}$ is smaller than $\delta$, then the distance between $f(x)$ and $y$ is smaller than $\epsilon$.
$\lim _{x \rightarrow x_{0}} f(x)=y \Leftrightarrow \forall \epsilon>0 \exists \delta>0| | x-x_{0}|<\delta \Rightarrow| f(x)-y \mid<\epsilon$

## $\forall$ (for all), $\exists$ (exists) and $\exists$ ! (exists only one)

For all $x$ in P
For any $x$ in P
For each $x$ with the property $P$
There exists at least one $x$ in P
For at least one $x$ in P
There exists at least one $x$ with the property P
There exists exactly one $x$ in P

```
\forallx,x\inP;}\forallx\in
\forallx,x\inP;}\forallx\in
\forallx,P(x)
\existsx,x\inP;}\existsx\in
\existsx,x\inP;}\exists\textrm{x}\in
\existsx,P(x)
\exists!x,x\inP; \exists!x\inP
```


## Example

For all real numbers
For all real numbers smaller than 4
There exists at least one real number
There exists at least one real number greater than 2 There exists a single real number such that ...
$\forall x \in \mathbb{R}$
$\forall x \in \mathbb{R}, x<4$
$\exists x \in \mathbb{R}$
$\exists x \in \mathbb{R}, x>2$
$\exists!x \in \mathbb{R} \mid \ldots$

## (such that, it is verified, verifying)

## Example

There must be people that read all newspapers everyday. Let $P$ be the set of all persons, let $N$ be the set of all newspapers, and let $D$ be the set of all days. Then, the previous sentence is formalized as
$\exists p \in P|\forall d \in D| \forall n \in N \mid$ p reads n on d .
Literal reading: There exist at least one person such that for all days and for all newspapers it is verified that p reads n on d .

## Example

Every day, there must be someone that reads all newspapers.
$\forall d \in D|\exists p \in P| \forall n \in N \mid$ p reads n on d .
Literal reading: For all days it is verified that there exists at least one person verifying that for all newspapers it is verified that p reads n on d .

## (such that, it is verified, verifying)

## Example

$\lim _{x \rightarrow x_{0}} f(x)=y \Leftrightarrow \forall \epsilon>0 \exists \delta>0| | x-x_{0}|<\delta \Rightarrow| f(x)-y \mid<\epsilon$
Literal reading: the limit of $f(x)$ when $x$ goes to $x_{0}$ is $y$ if and only if for any $\epsilon$ greater than 0 , there exists $\delta$ greater than 0 such that if $\left|x-x_{0}\right|<\delta$ is true, then $|f(x)-y|<\epsilon$ is also true.

## Example

Fermat-Wiles Theorem:
$\forall n \in \mathbb{Z}, n>2\left|\forall(x, y, z) \in \mathbb{R}^{3}, x^{n}+y^{n}=z^{n}\right| x y z=0$
Literal reading: For all integer numbers it is verified that for any real numbers $x$, $y z$ with the property $x^{n}+y^{n}=z^{n}$ it is verified that at least one of the three numbers is 0 .

## Negation of qualifiers

Let's say we state that all elements in a given set $S$ has a certain property ( $\forall x \in S \mid P(x)$ ). The negation of this statement is that there exists at least one element of $S$ that does not have that property $(\exists x \in S \mid \overline{P(x)})$.

Similarly, if we state that there exists at least one element in a given set $S$ that has a certain property $(\exists x \in S \mid P(x))$. The negation of this statement is that none of the elements of $S$ have that property ( $\forall x \in S \mid \overline{P(x)})$.

## Negation of qualifiers

## Example

In a previous example we had: There must be people that read all newspapers everyday. Its negation is

$$
\begin{aligned}
& \exists p \in P|\forall d \in D| \forall n \in N \mid \text { p reads n on d. } \\
& \forall p \in P|\forall d \in D| \forall n \in N \mid \text { p reads n on d. } \\
& \forall p \\
& \forall p \in P|\exists d \in D| \forall n \in N \mid \text { p reads n on d. } \\
& \forall p \\
& \forall p \in P|\exists d \in D| \exists n \in N \mid \overline{\text { p reads n on d. }}= \\
& \forall p \in P|\exists d \in D| \exists n \in N \mid \text { p does not read n on d. }
\end{aligned}
$$

That is, For everybody, there is at least one day and one paper, such that $p$ did not read $n$ on $d$.

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## Modus ponens

The following proofs follow a reasoning model called Modus ponens which is formally written as
$(A \cap(A \Rightarrow B)) \Rightarrow B$.
The intuitive meaning is that if $A$ is true and $A \Rightarrow B$, then $B$ is also true. Most proofs follow this way of reasoning. They can be performed in a forward way
$A \Rightarrow B_{1} \Rightarrow B_{2} \Rightarrow \ldots \Rightarrow B$
or in a backward way
$B \Leftarrow B_{n} \Leftarrow B_{n-1} \Leftarrow \ldots \Leftarrow A$.

## Forward proofs $(A \Rightarrow B ; \mathrm{B}$ is necessary for A$)$

## Example

Prove that the third power of an odd number is odd.
Proof
Let there be the following propositions:

$$
\begin{aligned}
& \mathrm{A}: x \text { is odd. } \\
& \mathrm{B}: x^{3} \text { is odd. }
\end{aligned}
$$

We need to prove that $A \Rightarrow B$ ( $B$ is necessary for $A$ ).
Proof $A \Rightarrow B$
Since $x$ is an odd number we can write $x=2 k+1$ for some integer number $k$. Then, $x^{3}=(2 k+1)^{3}=8 k^{3}+12 k^{2}+6 k+1=2\left(4 k^{3}+6 k^{2}+3 k\right)+1=2 k^{\prime}+1$. For $k^{\prime}=4 k^{3}+6 k^{2}+3 k$, which is another integer number. Therefore, $x^{3}$ is odd.

## Forward proofs $(A \Rightarrow B ; \mathrm{B}$ is necessary for A$)$

## Example

A necessary condition for a natural number to be a multiple of 360 is that it is a multiple of 3 and 120 .

## Proof

Let there be the following propositions:

> A: To be multiple of 360
> B: To be multiple of 3 and 120

We need to prove that $A \Rightarrow B$ ( $B$ is necessary for $A$ ).
Proof $A \Rightarrow B$
Let x be a multiple of $360(\mathrm{~A}) \Rightarrow$ There exists a natural number $k$ such that $x=360 \cdot k \Rightarrow x=120 \cdot 3 \cdot k$. From this factorization, it is obvious that $x$ is a multiple of 120 and a multiple of 3 (B).

## Forward proofs $(A \Leftarrow B$; B is sufficient for A )

## Example

A sufficient condition for a natural number to be a multiple of 360 is that it is a multiple of 3 and 120. Proof
Let there be the following propositions:
A: To be multiple of 360
B1: To be multiple of 3
B2: To be multiple of 120
B: $\mathrm{B} 1 \cap \mathrm{~B} 2$
We need to prove that $B \Rightarrow A$ ( $B$ is sufficient for $A$ ).
Proof $B \Rightarrow A$
We can easily prove that $B \nRightarrow A$ with a counterexample. Let us consider $x=240$. It is a multiple of 3 (B1). It is a multiple of 120 (B2). Therefore, $B$ is true. However, 240 is not a multiple of 360 (A is false). Therefore, we have proved that $B \nRightarrow A$.

## Forward proofs $(A \Rightarrow B)$

## Example

Show that $x=\frac{-b+\sqrt{b^{2}-4 a c}}{2 a}$ is solution of the equation $a x^{2}+b x+c=0$ Proof
Let there be the following propositions:

$$
\begin{aligned}
& A: x=\frac{-b+\sqrt{b^{2}-4 a c}}{2 a} \\
& B: a x^{2}+b x+c=0
\end{aligned}
$$

We need to prove that $A \Rightarrow B$.
If $A \Rightarrow B$ is true, then it must also be true that $A \Rightarrow B_{1}$

$$
B_{1}: a\left(\frac{-b+\sqrt{b^{2}-4 a c}}{2 a}\right)^{2}+b \frac{-b+\sqrt{b^{2}-4 a c}}{2 a}+c=0
$$

that we can rewrite as

$$
B_{1}: a\left(\frac{b^{2}}{4 a^{2}}+\frac{b^{2}-4 a c}{4 a^{2}}-\frac{2 b \sqrt{b^{2}-4 a c}}{4 a^{2}}\right)+b \frac{-b+\sqrt{b^{2}-4 a c}}{2 a}+c=0
$$

## Forward proofs $(A \Rightarrow B)$

## Example (continued)

that we can simplify to

$$
\begin{aligned}
& B_{1}: \frac{b^{2}}{4 a}+\frac{b^{2}}{4 a}-c-\frac{b \sqrt{b^{2}-4 a c}}{2 a}+\frac{-b^{2}}{2 a}+\frac{b \sqrt{b^{2}-4 a c}}{2 a}+c=0 \\
& B_{1}: \frac{b^{2}}{4 a}+\frac{b^{2}}{4 a}-\phi-\frac{b \sqrt{b^{2}-4 a c}}{2 a}+\frac{-b^{2}}{2 a}+\frac{b \sqrt{b^{2}-4 a c}}{2 a}+\phi=0 \\
& B_{1}: 0=0
\end{aligned}
$$

Since $B_{1}$ is always true (a statement that is always true is called a tautology), then $A \Rightarrow B_{1}$ is true, as we wanted.

## Forward proofs $(A \Leftrightarrow B)$

In this case we have to prove both directions: $A \Rightarrow B$ and $B \Rightarrow A$.

## Example

A necessary and sufficient condition for a natural number to be a multiple of 360 is that it is a multiple of 5 and 72 .
Proof
Let there be the following propositions:
A: To be multiple of 360
B1: To be multiple of 5
B2: To be multiple of 72
B: $\mathrm{B} 1 \cap \mathrm{~B} 2$
We need to prove that $A \Leftrightarrow B$, that is, $A \Rightarrow B$ and $B \Rightarrow A$
Proof $A \Rightarrow B$
Let $x$ be a multiple of $360(A) \Rightarrow$ There exists a natural number $k$ such that $x=360 \cdot k \Rightarrow x=72 \cdot 5 \cdot k$. From this factorization, it is obvious that $x$ is a multiple of 72 and a multiple of 5 (B).

## Forward proofs $(A \Leftrightarrow B)$

## Example (continued)

## Proof $B \Rightarrow A$

$B 1 \Rightarrow$ There is a natural number $k_{1}$ such that $x=5 \cdot k_{1}$ $B 2 \Rightarrow$ There is a natural number $k_{2}$ such that $x=72 \cdot k_{2}$ Therefore, $5 k_{1}=72 k_{2} \Rightarrow k_{1}=\frac{72}{5} k_{2}$. But $k_{1}$ is a natural number not a rational number, therefore, $k_{2}$ needs to be a multiple of 5 , i.e., there exists a natural number $k_{3}$ such that $k_{2}=5 \cdot k_{3}$. Consequently, considering B2, we have $x=72 \cdot 5 \cdot k_{3}=360 \cdot k_{3}$. That is $x$ is a multiple of 360 . Therefore, we have proved that $A \Rightarrow B$.

## More forward proofs

If I want to prove that A does not imply $\mathrm{B}(A \Rightarrow B$ is false $)$, I have to prove that $B$ is false, but $A$ is true.

## Example

In our example, I have to prove that you did not bring 1000 euros (B is false), but the bank is open ( A is true). I don't have to prove that

- B is false (you did not bring 1000 euros)
- A is false (the bank is closed)
- B is true but A is false (you brought 1000 euros, but the bank is closed)
- A and B are false (you did not bring 1000 euros, and the bank is closed)


## More forward proofs

If I know that B is false, and I want to proof that A implies $\mathrm{B}(A \Rightarrow B$ is true $)$, then I have to prove that $A$ is also false.

$$
\begin{array}{cc|c}
A & B & A \Rightarrow B \\
\hline \mathrm{~F} & \mathrm{~F} & \mathbf{T} \\
\mathrm{~F} & \mathrm{~T} & \mathrm{~T} \\
\mathrm{~T} & \mathrm{~F} & \mathrm{~F} \\
\mathrm{~T} & \mathrm{~T} & \mathrm{~T}
\end{array}
$$

## Example

If I know that you did not bring 1000 euros ( $B$ is false), all I have to prove to show that $A \Rightarrow B$ is true, is that the bank is closed ( A is false).

## More forward proofs

If I want to proof that A implies B or $\mathrm{C}(A \Rightarrow B \cup C$ is true $)$, and I prove that it is false that $A \Rightarrow B \cap C$, have I finished? No, let's see why

| $A$ | $B$ | $C$ | $B \cup C$ | $A \Rightarrow B \cup C$ | $B \cap C$ | $A \Rightarrow B \cap C$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| F | F | F | F | T | F | T |
| F | F | T | T | T | F | T |
| F | T | F | T | T | F | T |
| F | T | T | T | T | T | T |
| T | F | F | F | F | F | F |
| T | F | T | T | T | F | F |
| T | T | F | T | T | F | F |
| T | T | T | T | T | T | T |

## More forward proofs

If I prove that $A \Rightarrow B \cap C$ is false, that amounts to selecting the following rows from the table

| $A$ | $B$ | $C$ | $B \cup C$ | $A \Rightarrow B \cup C$ | $B \cap C$ | $A \Rightarrow B \cap C$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| T | F | F | F | F | F | F |
| T | F | T | T | T | F | F |
| T | T | F | T | T | F | F |

In those lines, $A \Rightarrow B \cup C$ is true for two of the $A, B, C$ combinations (that's good), but false for the other (that's bad). Therefore, we have not finished yet and we have to prove that either B or C is true, so that we can finally reduce the table to

$$
\begin{array}{ccc|c|c||c|c}
A & B & C & B \cup C & A \Rightarrow B \cup C & B \cap C & A \Rightarrow B \cap C \\
\hline \mathrm{~T} & \mathrm{~F} & \mathrm{~T} & \mathrm{~T} & \mathrm{~T} & \mathrm{~F} & \mathrm{~F} \\
\mathrm{~T} & \mathrm{~T} & \mathrm{~F} & \mathrm{~T} & \mathrm{~T} & \mathrm{~F} & \mathrm{~F}
\end{array}
$$

in which $A \Rightarrow B \cup C$ is true, and consequently, we have proved that $A \Rightarrow B \cup C$.

## Backward proofs $(A \Rightarrow B)$

## Example

Show that if $x>0$, then $x+\frac{1}{x} \geq 2$.
Proof
Let there be the following propositions:

$$
A: x>0
$$

$$
\text { B: } x+\frac{1}{x} \geq 2
$$

It is obvious that $C_{1} \Rightarrow B, C_{2} \Rightarrow C_{1}, C_{3} \Rightarrow C_{2}$ being
C1: $x+\frac{1}{x}-2 \geq 0$
C2: $\frac{x^{2}+1-2 x}{x} \geq 0$
C3: $\frac{(x-1)^{2}}{x} \geq 0$
It is also obvious that $A \Rightarrow C_{3}$ and, in this way, we have proved that $A \Rightarrow B$. We can simplify the writing of this proof as:

$$
x+\frac{1}{x} \geq 2 \Leftarrow x+\frac{1}{x}-2 \geq 0 \Leftarrow \frac{x^{2}+1-2 x}{x} \geq 0 \Leftarrow \frac{(x-1)^{2}}{x} \geq 0 \Leftarrow x>0
$$

## Backward proofs $(A \Rightarrow B)$

## Example

If $x, y \in \mathbb{R}, x, y>0$, then $\sqrt{x y} \leq \frac{x+y}{2}$
Proof
$\overline{\sqrt{x y}} \leq \frac{x+y}{2} \Leftarrow \sqrt{x y}-\frac{x+y}{2} \leq 0 \Leftarrow \frac{x+y}{2}-\sqrt{x y} \geq 0$
Since $x$ and $y$ are positive numbers, we can write them as $x=a^{2}$ and $y=b^{2}$.
Then, $\frac{x+y}{2}-\sqrt{x y} \geq 0 \Leftarrow \frac{a^{2}+b^{2}}{2}-a b \geq 0 \Leftarrow a^{2}+b^{2}-2 a b \geq 0 \Leftarrow(a-b)^{2} \geq 0$
This last proposition is always true, therefore $\sqrt{x y} \leq \frac{x+y}{2}$ is also true.

## Outline

(0) Mathematical language

- Axioms, postulates, definitions and propositions (a)
- Logical operators (a)
- Qualifiers (b)
- Mathematical proofs
- Modus ponens (b)
- Modus tollens (c)
- Reductio ad absurdum (c)
- Induction (c)
- Case distinction (c)
- Counterexample (c)
- Common math mistakes (c)


## Modus tollens

The following proofs follow a reasoning model called Modus tollens which is formally written as

$$
(\bar{B} \cap(A \Rightarrow B)) \Rightarrow \bar{A} .
$$

The intuitive meaning is that if $A \Rightarrow B$ is true and $B$ is false, then $A$ must also be false. Another way of writing this reasoning is
$(A \Rightarrow B) \Leftrightarrow(\bar{B} \Rightarrow \bar{A})$.
That is if we want to prove $A \Rightarrow B$, it is enough to prove $\bar{B} \Rightarrow \bar{A}$.

## Modus tollens

## Example

Show that if $x^{3}$ is even, then $x$ is even.
Proof
Let there be the following propositions:
A: $x^{3}$ is even
B: $x$ is even
We want to prove that $A \Rightarrow B$. Instead, we'll prove that $\bar{B} \Rightarrow \bar{A}$, with
$\bar{B}: x$ is odd
$\bar{A}: x^{3}$ is odd
But we already proved this in a previous example. Therefore, $A \Rightarrow B$ is true.

## Modus tollens

## Example

Show that if $c$ is odd, then the equation $n^{2}+n-c=0$ has no integer solution.
Proof
Let there be the following propositions:
A: $c$ is odd
B: $n^{2}+n-c=0$ has no integer solution
We want to prove that $A \Rightarrow B$. Instead, we'll prove that $\bar{B} \Rightarrow \bar{A}$, with $\bar{B}: n^{2}+n-c=0$ has an integer solution
$\bar{A}$ : $c$ is even
Proof $\bar{B} \Rightarrow \bar{A}$
Let's assume that $n \in \mathbb{Z}$ is solution of $n^{2}+n-c=0$. If $n$ is even, then $c$ is even because $c=n^{2}+n=(2 k)^{2}+2 k=2\left(2 k^{2}+k\right)$. If $n$ is odd, then $c$ is also even because $c=n^{2}+n=(2 k+1)^{2}+(2 k+1)=$ $2\left(2 k^{2}+3 k+1\right)$.

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- Common math mistakes (c)


## Reductio ad absurdum

The following proofs follow a reasoning model called Reductio ad absurdum which is formally written as
$A \Rightarrow B \Leftrightarrow(A \cap \bar{B} \Rightarrow$ absurdum $)$.
Absurdum is a statement that is always false, like $P \cap \bar{P}$. Let's analyze the truth table for this proposition

## Truth table

| $A$ | $B$ | $A \Rightarrow B$ | $A \cap \bar{B}$ | $P \cap \bar{P}$ | $A \cap \bar{B} \Rightarrow(P \cap \bar{P})$ |
| :---: | :---: | :---: | :---: | :---: | :---: |
| F | F | T | F | F | T |
| F | T | T | F | F | T |
| T | F | F | T | F | F |
| T | T | T | F | F | T |

We see that the third and sixth columns are identical.

## Reductio ad absurdum

## Example

Show that $\sqrt{2}$ is irrational.
Proof
It does not appear in the form $A \Rightarrow B$ but it can be put with
A: All facts we know about numbers
B: $\sqrt{2}$ is irrational
Let's assume that $\sqrt{2}$ is rational $(\bar{B})$, that is $\exists p, q \in \mathbb{Z} \left\lvert\, \sqrt{2}=\frac{p}{q}\right.$ and $p, q$ are irreducible (they don't have any common factor). If this is true, then $2 q^{2}=p^{2}$, i.e., 2 must be a factor of $p$ and consequently $p$ must be $p=2 r$. Substituting this knowledge into $2 q^{2}=p^{2}$ we obtain $2 q^{2}=(2 r)^{2} \Rightarrow q^{2}=2 r^{2}$. Consequently, 2 is another factor of $q$. But we presumed that
$P: p$ and $q$ were irreducible
So, if $\sqrt{2}$ is rational, then we have $P$ and $\bar{P}$ at the same time, which is a contradiction, and therefore $\sqrt{2}$ cannot be rational.

## Reductio ad absurdum

## Example

Show that there are infinite prime numbers.
Proof
Let's presume they is a finite list of prime numbers (in ascending order):
$2,3,5,7, \ldots, P$
Now we construct the number $M=2 \cdot 3 \cdot 5 \cdot 7 \cdot \ldots \cdot P+1$.
If $M$ is prime, then we have a contradiction is $M$ is prime and is larger than $P$.
If $M$ is not prime, then it has as a factor at least one of the prime numbers in the list. Let's assume it is 3 , that is
$M=3 H=2 \cdot 3 \cdot 5 \cdot 7 \cdot \ldots \cdot P+1 \Rightarrow 1=3(H-2 \cdot 5 \cdot 7 \cdot \ldots \cdot P)$
that means that 3 is a factor of 1 , which is an absurdum.

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## Weak induction

This is a strategy to prove a property of a natural number, $P(n)$. We follow the strategy below:
(1) Prove that $P(k)$ is true.
(2) Prove that if $P(n-1)$ is true, then $P(n)$ is also true

## Example

Show that $S_{n}=\sum_{i=1}^{n} i=\frac{n(n+1)}{2}$
Proof
(1) $S_{1}=\sum_{i=1}^{1} i=\frac{1(1+1)}{2}=1$, which is obviusly true.
(2) Let's assume that $S_{n-1}=\sum_{i=1}^{n-1} i=\frac{(n-1) n}{2}$. Then, we need to prove that

$$
\begin{aligned}
& S_{n}=\sum_{i=1}^{n} i=\frac{n(n+1)}{2} \text {. But } \\
& S_{n}=S_{n-1}+n=\frac{(n-1) n}{2}+n=n\left(\frac{n-1}{2}+1\right)=\frac{n(n+1)}{2} \text {. q.e.d. }
\end{aligned}
$$

## Strong induction

The goal is similar to the previous method, but now in the second step we assume that the property is true for all previous integers
(1) Prove that $P(k)$ is true.
(2) Prove that if $P(k)$ is true and $P(k+1)$ is true and $\ldots P(n-1)$ is true, then $P(n)$ is also true

## Example: Fundamental theorem of arithmetics

Show that for all natural numbers larger than 1 either it is prime or it is the product of prime numbers
Proof
(1) The property is true for 2 .
(2) Let's assume that it is true for $2,3,4, \ldots, n-1$.

If $n$ is prime, then the property is also true for $n$.
If $n$ is not prime, then it can be written as the product of several numbers between 2 and $n-1$. But the property is true for all these numbers, and therefore, the property is also true for $n$.

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## Case distinction

For each case we follow a different strategy.

## Example: Triangular inequality

Show that $\forall a, b \in \mathbb{R}| | a+b|\leq|a|+|b|$
Proof
We remind that the absolute value is a function defined by parts:
$|x|=\left\{\begin{array}{cc}x & x \geq 0 \\ -x & x<0\end{array}\right.$
Case $a+b \geq 0: a+b \leq|a|+|b|$
For all real numbers it is obvious that $x \leq|x|$. Therefore, we have $a \leq|a|$ and $b \leq|b|$. Consequently, $a+b \leq|a|+|b|$.
Case $a+b<0$ : $-(a+b) \leq|a|+|b|$
For all real numbers it is also true that $-x \leq|x|$. Therefore, we have
$-a \leq|a|$ and $-b \leq|b|$. Consequently, $-(a+b)=-a-b \leq|a|+|b|$.

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## Counterexample

To prove that something is not true, it is enough to show that it is not true for one example. This example is called a counterexample.

## Example

Show that $\forall x, y, z \in \mathbb{R}^{+}$and $\forall n \in \mathbb{Z}, n \geq 2$ it is verified that $x^{n}+y^{n} \neq z^{n}$ Proof
The proposition is false because, for instance, for $x=3, y=4, z=5$ and $n=2$ we have

$$
3^{2}+4^{2}=5^{2}
$$

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## Common math mistakes

Avoid some common mathematical mistakes (many of them, algebraic):

- Common math mistakes Video 1: http://www.youtube.com/watch?v=VHo_sfVdieM
- Common math mistakes PDF:
http://tutorial.math.lamar.edu/pdf/Common_Math_Errors.pdf
- Common math mistakes Video 2:
http://www. youtube.com/watch?v=qHSUU_q_2wA
- Common math mistakes Video 3:
http://www.youtube.com/watch?v=cTiuocJfyCs
- Common math mistakes Video 4:
http://www.youtube.com/watch?v=r5Yro2GdJ6w


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# Chapter 1. Vectors 

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## Outline

(1) Vectors

- Vectors and basic operations (a)
- Linear combination (a)
- Inner product or dot product (b)
- Norm, vector length and unit vectors (b)
- Distances and angles (b)
- Multiplication by matrices (b)


## References

## Introduction to LINEAR ALGEBRA



## GILBERT STRANG

G. Strang. Introduction to linear algebra (4th ed). Wellesley Cambridge Press (2009). Chapter 1.

## References


D. Lay. Linear algebra and its applications (3rd ed). Pearson (2006). Chapter 1.

## A little bit of history

Vectors were developed during the XIX ${ }^{\text {th }}$ century by mathematicians and physicists like Carl Friedrich Gauss (1799), William Rowan Hamilton (1837), and James Clerk Maxwell (1873), mostly as a tool to represent complex numbers, and later as a tool to perform geometrical reasoning. Their modern algebra was formalized by Josiah Willard Gibbs (1901), a university professor at Yale.


To know more about the history of vectors visit

- http:
//www.math.mcgill.ca/labute/courses/133f03/VectorHistory.html
- https://www.math.ucdavis.edu/~temple/MAT21D/

SUPPLEMENTARY-ARTICLES/Crowe_History-of-Vectors.pdf

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(1) Vectors

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## What is a vector?

## Definition 1.1

Informally, a vector is a collection of n numbers of the same type. We say it has $n$ components $(1,2, \ldots, n)$

We'll see that this definition is terribly simplistic since many other things (like functions, infinite sequences, etc.) can be vectors. But, for the time being, let's stick to this simple definition.

## Example

```
    (c}\begin{array}{c}{-1}\\{0}\\{1}\end{array})\in\mp@subsup{\mathbb{Z}}{}{3}\quad\mathrm{ is a collection of 3 integer numbers
    (\begin{array}{c}{-1.1}\\{1.1}\\{-1.1}\\{\sqrt{}{2}}\end{array})\in\mp@subsup{\mathbb{Q}}{}{2}}\mathrm{ ( is a collection of 2 rational numb
Matlab:
[-1.1; sqrt(2)]
```


## Transpose

We distinguish between column vectors (for instance $\mathbf{v}$ below) and row vectors $(\mathbf{w})$. In the first case, we say $\mathbf{v}$ is a $n \times 1$ vector, while in the second, we say $\mathbf{w}$ is a $1 \times n$ vector.

$$
\mathbf{v}=\left(\begin{array}{c}
v_{1} \\
v_{2} \\
\ldots \\
v_{n}
\end{array}\right) \text { and } \mathbf{w}=\left(w_{1} w_{2} \ldots w_{n}\right) .
$$

## Definition 1.2

The transpose is the operation that transforms a column vector into a row vector and viceversa.

## Example

$$
(-11)^{T}=\binom{-1}{1}
$$

Matlab:
$\left[\begin{array}{ll}-1 & 1\end{array}\right]$,

## Addition of vectors

## Definition 1.3

Given two vectors $\mathbf{v}=\left(\begin{array}{c}v_{1} \\ v_{2} \\ \ldots \\ v_{n}\end{array}\right)$ and $\mathbf{w}=\left(\begin{array}{c}w_{1} \\ w_{2} \\ \ldots \\ w_{n}\end{array}\right)$ the sum of these two vectors
is another vector defined as $\mathbf{v}+\mathbf{w}=\left(\begin{array}{c}v_{1}+w_{1} \\ v_{2}+w_{2} \\ \ldots \\ v_{n}+w_{n}\end{array}\right)$. Note that you can only add two column vectors or two row vectors, but not a column and a row vector.

## Example

$$
\binom{-1.1}{1.1}+\binom{-1.1}{\sqrt{2}}=\binom{-2.2}{1.1+\sqrt{2}}
$$

Matlab:

$$
[-1.1 ; 1.1]+[-1.1 ; \operatorname{sqrt}(2)]
$$

## Properties 1.1

Commutativity:

$$
\mathbf{v}+\mathbf{w}=\mathbf{w}+\mathbf{v}
$$

## Addition of vectors

## Example

$$
\binom{4}{2}+\binom{-1}{2}=\binom{3}{4}
$$



## Product by scalar

## Definition 1.4

Given a vector $\mathbf{v}$ and a scalar $c$, the multiplication of $c$ and $\mathbf{v}$ is defined as

$$
c \mathbf{v}=\left(\begin{array}{c}
c v_{1} \\
c v_{2} \\
\ldots \\
c v_{n}
\end{array}\right)
$$

## Example

$$
\begin{aligned}
& 2\binom{-1.1}{1.1}=\binom{-2.2}{2.2} \\
& -\binom{-1.1}{1.1}=\binom{1.1}{-1.1}
\end{aligned}
$$

Matlab:

$$
2 *[-1.1 ; 1.1]-[1.1 ; 1.1]
$$

## Product by scalar

## Example

$$
\mathbf{w}=\binom{-1}{2}
$$

What is the shape of all scaled vectors of the form $\mathbf{c w}$ ?
If $\boldsymbol{w}=\mathbf{0}$, then it is a single point (0). If $\mathbf{w} \neq \mathbf{0}$, then it is the straight line that passes through $\mathbf{0}$ and $\mathbf{w}$.


## Properties

For simplification we will present them as properties for $\mathbb{R}^{n}$, but they apply to all vector spaces. Given any three vectors $\mathbf{u}, \mathbf{v}, \mathbf{w} \in \mathbb{R}^{n}$ and any two scalars $c, d \in \mathbb{R}$, we have

## Vector operation properties

Regarding the sum of vectors:
(1) $\mathbf{u}+\mathbf{v}=\mathbf{v}+\mathbf{u}$ Commutativity
(2) $(\mathbf{u}+\mathbf{v})+\mathbf{w}=\mathbf{u}+(\mathbf{v}+\mathbf{w})$ Associativity
(3) $\mathbf{u}+\mathbf{0}=\mathbf{0}+\mathbf{u}=\mathbf{u}$ Existence of neutral element
(1) $\mathbf{u}+-\mathbf{u}=-\mathbf{u}+\mathbf{u}=\mathbf{0}$ Existence of symmetric element

Regarding the sum of vectors and scalar product:
(6) $c(\mathbf{u}+\mathbf{v})=c \mathbf{v}+c \mathbf{u}$ Distributivity with respect to the sum of vectors
(0) $(c+d) \mathbf{u}=c \mathbf{u}+d \mathbf{u}$ Distributivity with respect to the sum of scalars

Regarding the scalar product:
(1) $c(\mathrm{~d} \mathbf{u})=(c d) \mathbf{u}$ Associativity
(B) $\mathbf{1 u}=\mathbf{u}$ Existence of neutral element

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- Vectors and basic operations (a)
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## Linear combination

## Definition 2.1

Given a collection of $p$ scalars $\left(x_{i}, i=1,2, \ldots, p\right)$ and $p$ vectors $\left(\mathbf{v}_{i}\right)$, the linear combination of the $p$ vectors using the weights given by the $p$ scalars is defined as

$$
\sum_{i=1}^{p} x_{i} \mathbf{v}_{i}=x_{1} \mathbf{v}_{1}+x_{2} \mathbf{v}_{2}+\ldots+x_{p} \mathbf{v}_{p}
$$

## Example

$$
\frac{1}{2}\binom{-1}{1}-\frac{2}{3}\binom{2}{2}=\binom{-\frac{5}{6}}{-\frac{11}{6}}
$$

Matlab:
format rational
$-1 / 2 *[-1 ; 1]-2 / 3 *[2 ; 2]$

## Linear combination

## Example

A very basic model of the activity of neurons is

$$
\text { output }=f\left(\sum_{i} \text { weight }_{i} \text { input }_{i}\right)
$$

where $f(x)$ is a non-linear function. In fact, this is the model used in artificial neuron networks.


The human brain has in the order of $10^{11}$ neurons and about $10^{18}$ connections. See https://www. youtube.com/watch?v=zLp-edwiGUU.

## Linear combination

## Example

$$
\mathbf{v}=\binom{4}{2} \mathbf{w}=\binom{-1}{2}
$$



We may think of the weight coefficients as the "travelling" instructions. For instance, for the figure in the right, the instructions say: "Travel $\frac{1}{3}$ of $\mathbf{v}$ along $\mathbf{v}$, then travel $\frac{1}{2}$ of $\mathbf{w}$ along $\mathbf{w}$ ".

## Linear combination

## What is the shape of all linear combinations of the form $c \mathbf{v}+d \mathbf{w}$

If the two vectors are not collinear (i.e., $\mathbf{w} \neq k \mathbf{v}$ ), then it is the whole plane passing by $\mathbf{0}, \mathbf{v}$ and $\mathbf{w}$. We can think of it as the sum of all vectors belonging to the line $\overline{\mathbf{0} v}$ and $\overline{\mathbf{0 w}}$.

The plane generated by $\mathbf{v}$ and $\mathbf{w}$ is the set of all vectors that can be generated as a linear combination of both vectors.
$\Pi=\{\mathbf{r} \mid \mathbf{r}=c \mathbf{v}+d \mathbf{w} \forall c, d \in \mathbf{R}\}$


## Linear combination

The previous example prompts the following definition:

## Definition 2.2 (Spanned subspace)

The subspace spanned by the vectors $\mathbf{v}_{i}, i=1,2, \ldots, p$, is the set of all vectors that can be expressed as the linear combination of them. Formally,

$$
\left\langle\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{p}\right\rangle=\operatorname{Span}\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{p}\right\} \triangleq\left\{\mathbf{v} \in \mathbb{R}^{n} \mid \mathbf{v}=x_{1} \mathbf{v}_{1}+x_{2} \mathbf{v}_{2}+\ldots+x_{p} \mathbf{v}_{p}\right\}
$$

## Example

Assuming all vectors below are linearly independent:
$\operatorname{Span}\left\{\mathbf{v}_{1}\right\}$ is a straight line.
$\operatorname{Span}\left\{\mathbf{v}_{1}, \mathbf{v}_{2}\right\}$ is a plane.
$\operatorname{Span}\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{n-1}\right\}$ is a hyperplane.
$\operatorname{Span}\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{n-1}\right\}$ is a hyperplane.

## Properties

$$
\mathbf{0} \in \operatorname{Span}\{\cdot\}
$$

## Linear combination

## Outside the plane

Let $\mathbf{v}=(1,1,0)$ and $\mathbf{w}=(0,1,1)$. The linear combinations of $\mathbf{v}$ and $\mathbf{w}$ fill a plane in 3D. All points belonging to this plane are of the form
$\Pi=\{\mathbf{r} \mid \mathbf{r}=c(1,1,0)+d(0,1,1) \forall c, d \in \mathbf{R}\}=\{\mathbf{r}=(c, c+d, d) \forall c, d \in \mathbf{R}\}$
It is clear that the vector $\mathbf{r}^{\prime}=(0,1,0) \notin \Pi$, therefore, it is outside the plane.


## Linear combination

## Sets of points

Let $\mathbf{v}=(1,0)$.
(1) $S_{1}=\{\mathbf{r}=c \mathbf{v} \forall c \in \mathbb{Z}\}$ is a set of points
(2) $S_{2}=\left\{\mathbf{r}=c \mathbf{v} \forall c \in \mathbb{R}^{+}\right\}$is a semiline
(3) $S_{3}=\{\mathbf{r}=c \mathbf{v} \forall c \in \mathbb{R}\}$ is a line




## Linear combination

## Sets of points

Let $\mathbf{v}=(1,0)$ and $\mathbf{w}=(0,1)$.
(1) $S_{1}=\{\mathbf{r}=c \mathbf{v}+d \mathbf{w} \forall c \in \mathbb{Z}, \forall d \in \mathbb{R}\}$ is a set of lines
(2) $S_{2}=\left\{\mathbf{r}=c \mathbf{v}+d \mathbf{w} \forall c \in \mathbb{R}^{+}, \forall d \in \mathbb{R}\right\}$ is a semiplane
(0) $S_{3}=\{\mathbf{r}=c \mathbf{v}+d \mathbf{w} \forall c, d \in \mathbb{R}\}$ is a plane




## Linear combination

## Combination coefficients

Let $\mathbf{v}=(2,-1), \mathbf{w}=(-1,2)$ and $\mathbf{b}=(1,0)$. Find $c$ and $d$ such that $\mathbf{b}=c \mathbf{v}+d \mathbf{w}$.

## Solution

We need to find $c$ and $d$ such that

$$
\binom{1}{0}=c\binom{2}{-1}+d\binom{-1}{2}=\binom{2 c-d}{2 d-c}
$$

This gives a simple equation system

$$
\begin{aligned}
& 2 c-d=1 \\
& 2 d-c=0
\end{aligned}
$$

whose solution is $c=\frac{2}{3}$ and $d=\frac{1}{3}$. We can easily check it with Matlab:
$2 / 3 *\left[\begin{array}{ll}2 & -1\end{array}\right] '+1 / 3 *\left[\begin{array}{ll}-1 & 2\end{array}\right]$ '

## Exercises

## Exercises

From Lay (4th ed.), Chapter 1, Section 3:

- 1.3.1
- 1.3.3
- 1.3.6
- 1.3.7
- 1.3.25
- 1.3.27
- 1.3.29
- 1.3.31


## Outline

(1) Vectors

- Vectors and basic operations (a)
- Linear combination (a)
- Inner product or dot product (b)
- Norm, vector length and unit vectors (b)
- Distances and angles (b)
- Multiplication by matrices (b)


## Inner product

## Definition 3.1

Given two vectors $\mathbf{v}$ and $\mathbf{w}$ the inner or dot product between $\mathbf{v}$ and $\mathbf{w}$ is defined as

$$
\langle\mathbf{v}, \mathbf{w}\rangle=\mathbf{v} \cdot \mathbf{w} \triangleq \mathbf{v}^{T} \mathbf{w}=\sum_{i=1}^{n} v_{i} w_{i}=v_{1} w_{1}+v_{2} w_{2}+\ldots+v_{n} w_{n}
$$

Mathematically, the concept of inner product is much more general, and this operational definition is just a particularization for vectors in $\mathbb{R}^{n}$. Although, the introduced inner product is the most common, it is not the only one that can be defined in $\mathbb{R}^{n}$. But, let's leave these generalization for the moment.

Example

$$
\binom{4}{2} \cdot\binom{-1}{2}=4 \cdot(-1)+2 \cdot 2=0
$$

## Properties 3.1

Commutativity:
$\mathbf{v} \cdot \mathbf{w}=\mathbf{w} \cdot \mathbf{v}$

Matlab:

$$
\operatorname{dot}([4 ; 2],[-1 ; 2])
$$

## Outline

(1) Vectors

- Vectors and basic operations (a)
- Linear combination (a)
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## Vector norm and vector length

## Definition 4.1

Given a vector $\mathbf{v}$, its length or norm is defined as

$$
\|\mathbf{v}\| \triangleq \sqrt{\langle\mathbf{v}, \mathbf{v}\rangle}
$$

In the particular case of working with the previously introduced inner product, this definition boils down to

$$
\|\mathbf{v}\| \triangleq \sqrt{\mathbf{v}^{\top} \mathbf{v}}=\sqrt{\sum_{i=1}^{n} v_{i}^{2}}
$$

that is known as the Euclidean norm of vector $\mathbf{v}$.

## Properties 4.1

$$
\begin{aligned}
\|-\mathbf{v}\| & =\|\mathbf{v}\| \\
\|c \mathbf{v}\| & =\mid c\|\mathbf{v}\|
\end{aligned}
$$

## Vector norm and vector length

## Example

$$
\|(-1,0,1)\|=\sqrt{(-1)^{2}+0^{2}+1^{2}}=\sqrt{2}
$$

Matlab:

```
norm([-1;0;1])
```



## Unit vectors

## Definition 4.2

$\mathbf{v}$ is unitary iff $\|\mathbf{v}\|=1$.

## Example

$$
\begin{aligned}
& \mathbf{e}_{1}=(1,0) \\
& \mathbf{e}_{2}=(0,1) \\
& \mathbf{e}_{\theta}=(\cos (\theta), \sin (\theta)) \\
& \text { Matlab: } \\
& \quad \text { theta=pi/4; } \\
& \text { e_theta= } \left.^{\text {etos }} \text { (theta) } ; \sin (\text { theta })\right] ; \\
& \text { norm }\left(e_{-}\right. \text {theta) }
\end{aligned}
$$

## Unit vectors

## Definition 4.3 (Construction of a unit vector)

Given any vector v (whose norm is not null), we can always construct a unitary vector with the same direction of $\mathbf{v}$ as $\mathbf{u}_{\mathbf{v}}=\frac{\mathbf{v}}{\|\mathbf{v}\|}$.

## Example

$$
\begin{aligned}
& \mathbf{v}=(1,1) \\
& \mathbf{u}_{\mathbf{v}}=\frac{\mathbf{v}}{\|\mathbf{v}\|}=\frac{(1,1)}{\sqrt{2}}=\left(\frac{1}{\sqrt{2}}, \frac{1}{\sqrt{2}}\right)
\end{aligned}
$$



## Outline

(1) Vectors

- Vectors and basic operations (a)
- Linear combination (a)
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- Distances and angles (b)
- Multiplication by matrices (b)


## Distance and angle between two vectors

## Definition 5.1

Given two vectors $\mathbf{v}$ and $\mathbf{w}$, the distance between both is defined as

$$
d(\mathbf{v}, \mathbf{w}) \triangleq\|\mathbf{v}-\mathbf{w}\|
$$

and their angle is

$$
\angle(\mathbf{v}, \mathbf{w}) \triangleq \operatorname{acos} \frac{\mathbf{v} \cdot \mathbf{w}}{\|\mathbf{v}\|\|\mathbf{w}\|}=\theta
$$



## Definition 5.2

Two vectors are orthogonal (perpendicular) iff their inner product is 0 . We then write $\mathbf{v} \perp \mathbf{w}$. In this case, $\angle(\mathbf{v}, \mathbf{w})=\frac{\pi}{2}$.

## Distance and angle between two vectors

## Example

Let $\mathbf{v}=\left(-\frac{2}{5}, \frac{2}{3}\right)$ and $\mathbf{w}=\left(1, \frac{2}{3}\right)$. The angle between these two vectors can be calculated as

$$
\begin{aligned}
& \mathbf{v} \cdot \mathbf{w}=\left(-\frac{2}{5}\right) 1+\frac{2}{3} \frac{2}{3}=\frac{2}{45} \\
& \|\mathbf{v}\|=\sqrt{\left(-\frac{2}{5}\right)^{2}+\left(\frac{2}{3}\right)^{2}}=\frac{\sqrt{136}}{15} \\
& \|\mathbf{w}\|=\sqrt{(1)^{2}+\left(\frac{2}{3}\right)^{2}}=\frac{\sqrt{13}}{3} \\
& \angle(\mathbf{v}, \mathbf{w})=\operatorname{acos} \frac{\frac{2}{45}}{\frac{\sqrt{136}}{15} \frac{\sqrt{13}}{3}}=87.27^{\circ} \\
& \mathbf{v} \text { and } \mathbf{w} \text { are almost orthogonal. }
\end{aligned}
$$



## Example

Let $\mathbf{v}=(1,0,0,1,0,0,1,0,0,1)$ and $\mathbf{w}=(0,1,1,0,1,1,0,1,1,0)$. These two vectors in a 10 -dimensional space are orthogonal because
$\mathbf{v} \cdot \mathbf{w}=1 \cdot 0+0 \cdot 1+0 \cdot 1+1 \cdot 0+0 \cdot 1+0 \cdot 1+1 \cdot 0+0 \cdot 1+0 \cdot 1+1 \cdot 0=0$

## Distance and angle between two vectors

## Example

Search for a vector that is orthogonal to $\mathbf{v}=\left(-\frac{2}{5}, \frac{2}{3}\right)$
Solution
Let the vector $\mathbf{w}=\left(w_{1}, w_{2}\right)$ be such a vector. Since it is orthogonal to $\mathbf{v}$ it must meet

$$
\langle\mathbf{v}, \mathbf{w}\rangle=0=\left(-\frac{2}{5}\right) w_{1}+\frac{2}{3} w_{2} \Rightarrow w_{2}=\frac{3}{5} w_{1}
$$

That is, any vector of the form $\mathbf{w}=\left(w_{1}, \frac{3}{5} w_{1}\right)=w_{1}\left(1, \frac{3}{5}\right)$ is perpendicular to $\mathbf{v}$. This is the line passing by the origin and with direction $\left(1, \frac{3}{5}\right)$. In particular, for $w_{1}=\frac{2}{3}$ we have that $\mathbf{w}=\left(\frac{2}{3}, \frac{2}{5}\right)$ and for $w_{1}=-\frac{2}{3}$ we have $\mathbf{w}=\left(-\frac{2}{3},-\frac{2}{5}\right)$.

This is a general rule in 2D. Given a vector $\mathbf{v}=(a, b)$, the vectors $\mathbf{w}=(b,-a)$ and $\mathbf{w}=(-b, a)$ are orthogonal to $\mathbf{v}$.

$$
(a, b) \perp(b,-a) \text { and }(a, b) \perp(-b, a)
$$

## Distance and angle between two vectors

## Theorem 5.1 (Pythagorean theorem)

If $\mathbf{v} \perp \mathbf{w}$, then $\|\mathbf{v}-\mathbf{w}\|^{2}=\|\mathbf{v}\|^{2}+\|\mathbf{w}\|^{2}$.
Proof
$\|\mathbf{v}-\mathbf{w}\|^{2}=(\mathbf{v}-\mathbf{w})^{T}(\mathbf{v}-\mathbf{w})=\mathbf{v}^{T} \mathbf{v}-\mathbf{v}^{T} \mathbf{w}-\mathbf{w}^{T} \mathbf{v}+\mathbf{w}^{T} \mathbf{w}=\|\mathbf{v}\|^{2}+\|\mathbf{w}\|^{2}-2\langle\mathbf{v}, \mathbf{w}\rangle$
But, because $\mathbf{v} \perp \mathbf{w}$, we have $\langle\mathbf{v}, \mathbf{w}\rangle=0$, and consequently
$\|\mathbf{v}-\mathbf{w}\|^{2}=\|\mathbf{v}\|^{2}+\|\mathbf{w}\|^{2}$ (q.e.d.)

## Corollary 5.1

- If $\langle\mathbf{v}, \mathbf{w}\rangle<0$, then $\frac{\pi}{2}<\theta \leq \pi$.
- If $\langle\mathbf{v}, \mathbf{w}\rangle>0$, then $0 \leq \theta<\frac{\pi}{2}$.
- For two unit vectors, $\mathbf{u}_{1}$ and $\mathbf{u}_{2}$, we have $\cos \theta=\left\langle\mathbf{u}_{1}, \mathbf{u}_{2}\right\rangle$, and as a consequence $-1 \leq\left\langle\mathbf{u}_{1}, \mathbf{u}_{2}\right\rangle \leq 1$.


## Distance and angle between two vectors

## Theorem 5.2 (Cosine formula)

For any two vectors, $\mathbf{v}$ and $\mathbf{w}$, such that $\|\mathbf{v}\| \neq 0$ and $\|\mathbf{w}\| \neq 0$, we have

$$
\langle\mathbf{v}, \mathbf{w}\rangle=\|\mathbf{v}\|\|\mathbf{w}\| \cos \theta
$$

Proof
By use of Definition 4.3, we can construct the unit vectors associated to $\mathbf{v}$ and $\mathbf{w}$, that is $\mathbf{u}_{\mathbf{v}}$ and $\mathbf{u}_{\mathbf{w}}$. Then by Corollary 5.1 we know that

$$
\cos \theta=\left\langle\mathbf{u}_{\mathbf{v}}, \mathbf{u}_{\mathbf{w}}\right\rangle=\left(\frac{\mathbf{v}}{\|\mathbf{v}\|}\right)^{T}\left(\frac{\mathbf{w}}{\|\boldsymbol{w}\|}\right)=\frac{1}{\|\mathbf{u}\|\|\mathbf{w}\|} \mathbf{u}^{T} \mathbf{w}=\frac{\langle\mathbf{v}, \mathbf{w}\rangle}{\|\mathbf{u}\|\|\mathbf{w}\|}
$$

From this point it is trivial to deduce that $\langle\mathbf{v}, \mathbf{w}\rangle=\|\mathbf{v}\|\|\mathbf{w}\| \cos \theta$ (q.e.d.)

## Distance and angle between two vectors

## Example

To compute the knee flexion angle, we need to calculate the dot product between the vectors aligned with the leg before and after the knee.


## Distance and angle between two vectors

## Theorem 5.3 (Cauchy-Schwarz inequality)

For any two vectors, $\mathbf{v}$ and $\mathbf{w}$, it is verified that

$$
|\langle\mathbf{v}, \mathbf{w}\rangle|<\|\mathbf{v}\|\|\mathbf{w}\|
$$

Proof
From the cosine formula (Theorem 5.2), we know that

$$
\begin{aligned}
\langle\mathbf{v}, \mathbf{w}\rangle & =\|\mathbf{v}\|\|\mathbf{w}\| \cos \theta \Rightarrow \\
|\langle\mathbf{v}, \mathbf{w}\rangle| & =\|\mathbf{v}\|\|\mathbf{w}\| \cos \theta|=\|\mathbf{v}\|\|\mathbf{w}\|| \cos \theta \mid \leq\|\mathbf{v}\|\|\mathbf{w}\|
\end{aligned}
$$

## Example

Let $\mathbf{v}=\left(-\frac{2}{5}, \frac{2}{3}\right)$ and $\mathbf{w}=\left(1, \frac{2}{3}\right)$. We already know that $\mathbf{v} \cdot \mathbf{w}=\frac{2}{45},\|\mathbf{v}\|=\frac{\sqrt{136}}{15}$,
and $\|\mathbf{w}\|=\frac{\sqrt{13}}{3}$. Let us check Cauchy-Schwarz inequality

$$
\left|\frac{2}{45}\right|<\frac{\sqrt{136}}{15} \frac{\sqrt{13}}{3} \Leftrightarrow 0.0444<0.9344
$$

## Distance and angle between two vectors

## Example

Show that for any two positive numbers, $x$ and $y$, the geometric mean $(\sqrt{x y})$ is always smaller or equal than the arithmetic mean $\left(\frac{x+y}{2}\right)$. For instance, the statement is verified for $x=2$ and $y=3: \sqrt{6} \leq \frac{5}{2} \Leftrightarrow 2.4495 \leq 2.5$.
Proof
Let there be vectors $\mathbf{v}=(a, b)$ and $\mathbf{w}=(b, a)$. Then, by Cauchy-Schwarz inequality we know that

$$
|\langle\mathbf{v}, \mathbf{w}\rangle|<\|\mathbf{v}\|\|\mathbf{w}\| \Rightarrow|2 a b| \leq a^{2}+b^{2}
$$

Since $x$ and $y$ are positive numbers, we may consider them to be $x=a^{2}$ and $y=b^{2}$. Consequently, we can rewrite the previous expression as

$$
2 \sqrt{x} \sqrt{y} \leq x+y \Rightarrow \sqrt{x y} \leq \frac{x+y}{2} \text { (q.e.d.) }
$$

In fact, the geometric mean is nothing more than the arithmetic mean in logarithmic units

$$
\log (\sqrt{x y})=\log (x y)^{\frac{1}{2}}=\frac{1}{2}(\log x+\log y)=\frac{\log x+\log y}{2}
$$

## Distance and angle between two vectors

## Theorem 5.4 (Triangular inequality)

For any two vectors, $\mathbf{v}$ and $\mathbf{w}$, it is verified that

$$
\|\mathbf{v}+\mathbf{w}\| \leq\|\mathbf{v}\|+\|\mathbf{w}\|
$$

## Proof

By definition we know that

$$
\|\mathbf{v}+\mathbf{w}\|^{2}=(\mathbf{v}+\mathbf{w})^{T}(\mathbf{v}+\mathbf{w})=\|\mathbf{v}\|^{2}+\|\mathbf{w}\|^{2}+2\langle\mathbf{v}, \mathbf{w}\rangle
$$

Applying the Cauchy-Schwarz inequality (Theorem 5.3), we have

$$
\|\mathbf{v}+\mathbf{w}\|^{2} \leq\|\mathbf{v}\|^{2}+\|\mathbf{w}\|^{2}+2\|\mathbf{v}\|\|\mathbf{w}\|=(\|\mathbf{v}\|+\|\mathbf{w}\|)^{2}
$$

Taking the square root we have

$$
\|\mathbf{v}+\mathbf{w}\| \leq\|\mathbf{v}\|+\|\mathbf{w}\|
$$

## Distance and angle between two vectors

## Example

Let $\mathbf{v}=\left(-\frac{2}{5}, \frac{2}{3}\right)$ and $\mathbf{w}=\left(1, \frac{2}{3}\right)$. We already know that $\|\mathbf{v}\|=\frac{\sqrt{136}}{15}$ and $\|\mathbf{w}\|=\frac{\sqrt{13}}{3}$. Let us check the triangular inequality

$$
\begin{aligned}
\mathbf{v}+\mathbf{w} & =\left(\frac{3}{5}, \frac{4}{3}\right) \Rightarrow\|\mathbf{v}+\mathbf{w}\|=\frac{\sqrt{481}}{15} \\
\frac{\sqrt{481}}{15} & \leq \frac{\sqrt{136}}{15}+\frac{\sqrt{13}}{3} \Leftrightarrow 1.4621 \leq 1.9793
\end{aligned}
$$



## Distance and angle between two vectors

## Orthogonal projections

Let us consider the orthogonal projection of $\mathbf{v}$ onto $\mathbf{w}$.

$$
\mathbf{v}^{\prime}=\langle\mathbf{v}, \mathbf{w}\rangle \frac{\mathbf{w}}{\|\mathbf{w}\|^{2}}=\frac{\langle\mathbf{v}, \mathbf{w}\rangle}{\|\mathbf{w}\|} \frac{\mathbf{w}}{\|\mathbf{w}\|}
$$

The length of this vector is $\frac{\langle\mathbf{v}, \mathbf{w}\rangle}{\|\mathbf{w}\|}$


## Example

Let $\mathbf{v}=\left(\frac{5}{2}, 1\right)$ and $\mathbf{w}=(3,0)$. Then, $\mathbf{v}^{\prime}=\frac{\frac{5}{3} 3+1 \cdot 0}{3}(1,0)=\left(\frac{5}{2}, 0\right)$. See the figure above.

## Outline

(1) Vectors

- Vectors and basic operations (a)
- Linear combination (a)
- Inner product or dot product (b)
- Norm, vector length and unit vectors (b)
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- Multiplication by matrices (b)


## Multiplication by matrices

## Example

Let's consider three vectors $\mathbf{v}_{1}=\left(\begin{array}{c}1 \\ -1 \\ 0\end{array}\right), \mathbf{v}_{2}=\left(\begin{array}{c}0 \\ 1 \\ -1\end{array}\right)$ and $\mathbf{v}_{3}=\left(\begin{array}{c}0 \\ 0 \\ 1\end{array}\right)$. Let's consider the linear combination

$$
\mathbf{y}=x_{1} \mathbf{v}_{1}+x_{2} \mathbf{v}_{2}+x_{3} \mathbf{v}_{3}=x_{1}\left(\begin{array}{c}
1 \\
-1 \\
0
\end{array}\right)+x_{2}\left(\begin{array}{c}
0 \\
1 \\
-1
\end{array}\right)+x_{3}\left(\begin{array}{l}
0 \\
0 \\
1
\end{array}\right)=\left(\begin{array}{c}
x_{2} \\
x_{2}-x_{1} \\
x_{3}-x_{2}
\end{array}\right)
$$

I can obtain the same result by constructing a matrix

$$
A=\left(\begin{array}{lll}
\mathbf{v}_{1} & \mathbf{v}_{2} & \mathbf{v}_{3}
\end{array}\right)=\left(\begin{array}{ccc}
1 & 0 & 0 \\
-1 & 1 & 0 \\
0 & -1 & 1
\end{array}\right) .
$$

And making the multiplication

$$
\begin{aligned}
\mathbf{y}=A\left(\begin{array}{c}
x_{1} \\
x_{2} \\
x_{3}
\end{array}\right)=\left(\begin{array}{lll}
\mathbf{v}_{1} & \mathbf{v}_{2} & \mathbf{v}_{3}
\end{array}\right) & \left(\begin{array}{l}
x_{1} \\
x_{2} \\
x_{3}
\end{array}\right)=\left(\begin{array}{ccc}
1 & 0 & 0 \\
-1 & 1 & 0 \\
0 & -1 & 1
\end{array}\right)\left(\begin{array}{l}
x_{1} \\
x_{2} \\
x_{3}
\end{array}\right)= \\
& \left(\begin{array}{c}
x_{1} \\
x_{2}-x_{1} \\
x_{3}-x_{2}
\end{array}\right)
\end{aligned}
$$

## Multiplication by matrices

## Example

We can also achieve the same result by calculating $y$ as the inner product of the rows of the matrix $A$ and the weight vector.

$$
\mathbf{y}=\left(\begin{array}{c}
\left\langle(1,0,0),\left(x_{1}, x_{2}, x_{3}\right)\right\rangle \\
\left\langle(-1,1,0),\left(x_{1}, x_{2}, x_{3}\right)\right\rangle \\
\left\langle(0,-1,1),\left(x_{1}, x_{2}, x_{3}\right)\right\rangle
\end{array}\right)=\left(\begin{array}{c}
x_{2} \\
x_{2}-x_{1} \\
x_{3}-x_{2}
\end{array}\right)
$$

Matlab:

```
    syms x1 x2 x3
x=[x1; x2; x3]
A=[1 0 0; -1 1 0; 0 -1 1];
y=A*x
```


## Multiplication by matrices

## Matrix multiplication as a linear combination

This is a general rule: a matrix multiplication can be seen as the linear combination of the columns of the matrix.

$$
A=\left(\mathbf{c}_{1} \mathbf{c}_{2} \ldots \mathbf{c}_{p}\right) \Rightarrow \mathbf{y}=A \mathbf{x}=\sum_{i=1}^{p} x_{i} \mathbf{c}_{i}
$$

## Matrix multiplication as inner products

Also, a matrix multiplication can be seen as the dot product of the weight vector with the rows of the matrix.

$$
A=\left(\begin{array}{c}
\mathbf{r}_{1}^{T} \\
\mathbf{r}_{2}^{T} \\
\ldots \\
\cdots \\
\mathbf{r}_{n}^{T}
\end{array}\right) \Rightarrow \mathbf{y}=A \mathbf{x}=\left(\begin{array}{c}
\left\langle\mathbf{r}_{1}, \mathbf{x}\right\rangle \\
\left\langle\mathbf{r}_{2}, \mathbf{x}\right\rangle \\
\ldots \\
\left\langle\mathbf{r}_{n}, \mathbf{x}\right\rangle
\end{array}\right)
$$

## Multiplication by matrices

Properties of multiplication by matrices

$$
\begin{aligned}
A(\mathbf{u}+\mathbf{v}) & =A \mathbf{u}+A \mathbf{v} \\
A(c \mathbf{u}) & =c(A \mathbf{u})
\end{aligned}
$$

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(1) Vectors

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# Chapter 2. Linear equation systems 

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## Outline

(2) Linear equation system

- Introduction (a)
- Gauss-Jordan algorithm (b)
- Interpretation as a subspace (b)
- Existence and uniqueness of solutions (c)
- Applications (c)
- Linear independence (c)
- Linear transformations (d)
- Geometrical transformations (e)
- Classification of functions (e)
- More applications (e)


## References


D. Lay. Linear algebra and its applications (3rd ed). Pearson (2006). Chapter 1.

## A little bit of history

Linear equations in their modern form are known since the middle of the XVIII ${ }^{\text {th }}$ century and they were strongly developed during the XIX ${ }^{\text {th }}$ century with important contributions of people like Gabriel Cramer (1750), Carl Friedrich Gauss (1801), Sir William Rowan Hamilton (1843) and Wilhelm Jordan (1873). They were mostly developed to explain the mechanics of celestial objects.


To know more about the history of linear equations visit

- http://hom.wikidot.com/cramer-s-method-and-cramer-s-paradox


## A little bit of history

Wassily Leontief was a Russian-American economist that worked in Harvard. In 1949 he performed an analysis with the early computers at Harvard using data from the U.S. Bureau of Labor Statistics to classify the U.S. economy into 500 sectors, that were later simplified to 42 . He used linear equation systems to do so. It took 56 hours in Mark II (one of the first computers) to solve it. He was awarded the Nobel prize in 1970 for his work on input-output tables that analyze how outputs from some industries are inputs to some other industries.


## A little bit of history

Currently, we need about two weeks in a supercomputer (128 cores) to solve the structure of a macromolecular assembly (in the figure, the HIV virus capsid). We have 1,000 million equations with about 3 million unknowns.


## Outline

(2) Linear equation system

- Introduction (a)
- Gauss-Jordan algorithm (b)
- Interpretation as a subspace (b)
- Existence and uniqueness of solutions (c)
- Applications (c)
- Linear independence (c)
- Linear transformations (d)
- Geometrical transformations (e)
- Classification of functions (e)
- More applications (e)


## What is a linear equation system?

## Definition 1.1 (Linear equation system)

A linear equation is one that can be expressed in the form

$$
\begin{gathered}
\sum_{i=1}^{n} a_{i} x_{i}=b \\
a_{1} x_{1}+a_{2} x_{2}+\ldots+a_{n} x_{n}=b \\
\langle\mathbf{a}, \mathbf{x}\rangle=b
\end{gathered}
$$

The unknowns are $x_{i}(i=1,2, \ldots, n)$ while $a_{i}$ 's and $b$ are coefficients. When we have several of these equations, we have a linear equation system.

## Example

Examples of linear equations

$$
\begin{gathered}
7 x_{1}-2 x_{2}=4 \\
7\left(x_{1}-\sqrt{3} x_{2}\right)=\frac{1}{\sqrt{2}} x_{1} \Rightarrow \\
\left(7-\frac{1}{\sqrt{2}}\right) x_{1}-7 \sqrt{3} x_{2}=0
\end{gathered}
$$

## Examples of non-linear equations

$$
\begin{gathered}
x_{1}+x_{2}+x_{1} x_{2}=1 \\
\sqrt{x_{1}}+x_{2}=1
\end{gathered}
$$

## Set of solutions of a linear equation

## Definition 1.2 (Set of solutions of a linear equation system)

The set of solutions of a linear equation system $S \subseteq \mathbb{R}^{n}$ is the set of all those values that we can assign to $x_{1}, x_{2}, \ldots, x_{n}$ such that the equation system is fulfilled.

## Example

Consider the following equation system

$$
\begin{gathered}
2 x_{1}-x_{2}=7 \\
x_{1}+2 x_{2}=11
\end{gathered}
$$

$\mathbf{x}=(5,3)$ is a solution to this equation system because

$$
\begin{gathered}
2 \cdot 5-3=7 \\
5+2 \cdot 3=11
\end{gathered}
$$

In fact it is its unique solution and, therefore, $S=\{(5,3)\} \subset \mathbb{R}^{2}$.

## Geometric interpretation

## Example

$$
\begin{aligned}
& I_{1}: 2 x_{1}-x_{2}=7 \Rightarrow x_{2}=2 x_{1}-7 \Rightarrow \mathbf{v}_{1}=(1,2) \\
& I_{2}: x_{1}+2 x_{2}=11 \Rightarrow x_{2}=11-\frac{1}{2} x_{1} \Rightarrow \mathbf{v}_{2}=\left(1,-\frac{1}{2}\right)
\end{aligned}
$$

Each one of the equations is actually representing a line, and both lines, in this case intersect at the point $(5,3)$, the unique solution of this equation system.


## Geometric interpretation

## Example

There can be a single solution (left), no solution (middle), or infinite ( $I_{1}=I_{2}$; right)


## In general

With linear equations we can represent:
a line in 2D: $a_{1} x_{1}+a_{2} x_{2}=b$
a plane in 3D: $a_{1} x_{1}+a_{2} x_{2}+a_{3} x_{3}=b$
a hyperplane in $n \mathrm{D}: a_{1} x_{1}+a_{2} x_{2}+\ldots+a_{n} x_{n}=b$

## Matrix notation

## Example

The equation system

$$
\begin{aligned}
x_{1}-2 x_{2}+x_{3} & =0 \\
2 x_{2}-8 x_{3} & =8 \\
-4 x_{1}+5 x_{2}+9 x_{3} & =-9
\end{aligned}
$$

can be represented as

$$
\left(\begin{array}{rrr|r}
1 & -2 & 1 & 0 \\
0 & 2 & -8 & 8 \\
-4 & 5 & 9 & -9
\end{array}\right)[\tilde{A}]
$$

or

$$
\left(\begin{array}{rrr}
1 & -2 & 1 \\
0 & 2 & -8 \\
-4 & 5 & 9
\end{array}\right)\left(\begin{array}{l}
x_{1} \\
x_{2} \\
x_{3}
\end{array}\right)=\left(\begin{array}{r}
0 \\
8 \\
-9
\end{array}\right)[A \mathbf{x}=\mathbf{b}]
$$

## Matrix notation

## In general

$A \in \mathcal{M}_{m \times n}$ is called the system matrix of an equation system with $m$ equations and $n$ unknowns.
$\tilde{A} \in \mathcal{M}_{m \times(n+1)}$ is called the augmented system matrix of an equation system with $m$ equations and $n$ unknowns.

## Basic row iterations

To solve the equation system with the augmented system matrix, we used the so-called basic row operations:
Substitution: $\mathbf{r}_{i} \leftarrow k_{i} \mathbf{r}_{i}+k_{j} \mathbf{r}_{j}$ : Row $i$ is substituted by a linear combination of rows $i$ and $j$
Swapping: $\mathbf{r}_{i} \leftrightarrow \mathbf{r}_{j}$ : Row $i$ swapped with row $j$
Scaling: $\mathbf{r}_{i} \leftarrow k_{i} \mathbf{r}_{i}$ : Row $i$ is multiplied by a scale factor
All these operations transform the equation system into an equivalent system (with the same set of solutions). The two matrices (original and transformed) are said to be row equivalent.

## Solving the equation system

## Example

In the following example we will see how linear combinations are actually changing the equation system to a different one, while scaling is not.
PX1

## Solving the equation system

## Example

$$
\left(\begin{array}{rr|r}
1 & -\frac{1}{2} & \frac{7}{2} \\
1 & 2 & 11
\end{array}\right)
$$



$$
\mathbf{r}_{2} \leftarrow \mathbf{r}_{2}-\mathbf{r}_{1}
$$



## Solving the equation system

## Example

$$
\left(\begin{array}{rr|r}
1 & -\frac{1}{2} & \frac{7}{2} \\
0 & \frac{5}{2} & \frac{15}{2}
\end{array}\right)
$$



$$
\mathbf{r}_{2} \leftarrow \frac{2}{5} \mathbf{r}_{2}
$$



## Solving the equation system

## Example

$$
\left(\begin{array}{rr|r}
1 & -\frac{1}{2} & \frac{7}{2} \\
0 & 1 & 3
\end{array}\right)
$$

$$
\mathbf{r}_{1} \leftarrow \mathbf{r}_{1}+\frac{1}{2} \mathbf{r}_{2}
$$




## Existence and uniqueness of solutions

Example

$$
\begin{array}{rr}
x_{1}-2 x_{2}+x_{3} & =0 \\
& 2 x_{2}-8 x_{3}
\end{array}=8 \sim \ldots \sim\left(\begin{array}{rrr|r}
1 & -2 & 1 & 0 \\
0 & 1 & -4 & 4 \\
0 & 0 & 1 & 3
\end{array}\right)
$$

I can solve for $x_{3}\left(x_{3}=3\right)$, then use this value in the second equation to solve for $x_{2}$, and finally use these two values in the first equation to solve for $x_{1}$. Thus, the equation system has a solution and it is unique. We say the equation system is compatible. The set of solutions is $S=\{(29,16,3)\}$.

Matlab:

$$
\begin{aligned}
& \mathrm{A}=\left[\begin{array}{cccccccc}
1 & -2 & 1 ; & 0 & 2 & -8 ; & -4 & 5
\end{array}\right] ; \\
& \mathrm{b}=[0 ; 8 ; \\
& \mathrm{x}=\mathrm{A} \backslash \mathrm{~b}
\end{aligned}
$$

## Existence and uniqueness of solutions

## Example

$$
\begin{array}{rr}
x_{2}-4 x_{3} & =8 \\
2 x_{1} & -3 x_{2}+2 x_{3}
\end{array}=1 \sim \ldots \sim\left(\begin{array}{rrr|r}
2 & -3 & 2 & 1 \\
0 & 1 & -4 & 8 \\
5 x_{1} & -8 x_{2}+7 x_{3} & =1
\end{array}\right)
$$

Last equation implies $0=\frac{5}{2}$ which is impossible. Consequently, there is no solution and we say that the equation system is incompatible. The set of solutions is $S=\varnothing$.

## Example

$$
\begin{array}{r}
x_{1}+x_{2}=1 \\
2 x_{1}+2 x_{2}=2
\end{array} \sim \ldots \sim\left(\begin{array}{ll|l}
1 & 1 & 1 \\
0 & 0 & 0
\end{array}\right)
$$

There are infinite solutions. The system is compatible indeterminate. The set of solutions is $S=\left\{\left(x_{1}, 1-x_{1}\right)\right\}$.

## Exercises

## Exercises

From Lay (4th ed.), Chapter 1, Section 1:

- 1.1.11
- 1.1.4
- 1.1.15
- 1.1.18
- 1.1.25
- 1.1.26
- 1.1.33


## Outline

(2) Linear equation system

- Introduction (a)
- Gauss-Jordan algorithm (b)
- Interpretation as a subspace (b)
- Existence and uniqueness of solutions (c)
- Applications (c)
- Linear independence (c)
- Linear transformations (d)
- Geometrical transformations (e)
- Classification of functions (e)
- More applications (e)


## Echelon matrices

## Example

The following matrices are echelon matrices:

$$
\begin{aligned}
& A_{1}=\left(\begin{array}{llll}
\diamond & \diamond & \odot & \diamond \\
0 & \diamond & \diamond & \diamond \\
0 & 0 & 0 & 0 \\
0 & \diamond & \diamond & \diamond \\
0 & 0 & 0 & \diamond
\end{array}\right) \\
& A_{2}=\left(\begin{array}{ll} 
&
\end{array}\right)
\end{aligned}
$$

In the previous matrices we have marked with $\diamond$ the leading elements (the first ones different from 0 in their row), and with $\bigcirc$ the rest of the elements different from 0 .

## Echelon matrices

## Definition 2.1 (Echelon matrix)

A rectangular matrix has an echelon form iif:
(1) Within each row, the first element different from zero (called the leading entry) is in a column to the right of the leading entry of the previous row.
(2) Within each column, all values below a leading entry are zero.
(3) All rows without a leading entry (i.e., they only have zeros) are below all the rows in which at least one element is not zero.

## Definition 2.2 (Reduced echelon matrix)

A rectangular matrix has a reduced echelon form iif:
(1) It is echelon.
(2) The leading entry of each row is 1 .
(3) The leading entry is the only 1 in its column.

## Echelon matrices

## Theorem 2.1

Each matrix is row equivalent to one and only one reduced echelon matrix.

## Example

$$
\left.\begin{array}{l}
\mathbf{r}_{2} \leftarrow \mathbf{r}_{2}-4 \mathbf{r}_{1} \\
\mathbf{r}_{3} \leftarrow \mathbf{r}_{3}+\mathbf{r}_{1} \\
\mathbf{r}_{2} \leftrightarrow \mathbf{r}_{3} \\
\mathbf{r}_{3} \leftrightarrow \mathbf{r}_{3}+3 \mathbf{r}_{2} \left\lvert\,\left(\begin{array}{rrr}
1 & 2 & 3 \\
4 & 5 & 6 \\
-1 & -1 & 0 \\
1 & 2 & 3 \\
0 & -3 & -6 \\
0 & 1 & 3 \\
1 & 2 & 3 \\
0 & 1 & 3 \\
0 & -3 & -6
\end{array}\right)\right. \\
\left(\begin{array}{l}
\mathbf{r}_{1} \leftarrow \mathbf{r}_{1}-2 \mathbf{r}_{2} \\
\mathbf{r}_{3} \leftarrow \frac{1}{3} \mathbf{r}_{3} \\
1
\end{array}\left|\begin{array}{l}
\mathbf{r}_{1} \leftarrow \mathbf{r}_{1}+3 \mathbf{r}_{3} \\
\mathbf{r}_{2} \leftarrow \mathbf{r}_{2}-3 \mathbf{r}_{3}
\end{array}\right|\left(\begin{array}{ccc}
1 & 0 & -3 \\
0 & 1 & 3 \\
0 & 0 & 1 \\
1 & 0 & 0 \\
0 & 1 & 3 \\
0 & 0 & 1
\end{array}\right)\right.
\end{array}\right)
$$

## Echelon matrices

## Example (continued)

## Matlab:

$$
\begin{aligned}
& A=[123 ; 456 ;-1-10] \\
& A(2,:)=A(2,:)-4 * A(1,:) \\
& A(3,:)=A(3,:)+A(1,:) \\
& \operatorname{aux}=A(2,:) ; A(2,:)=A(3,:) ; A(3,:)=\text { aux } \\
& A(3,:)=A(3,:)+3 * A(2,:) \\
& A(1,:)=A(1,:)-2 * A(2,:) \\
& A(3,:)=1 / 3 * A(3,:) \\
& A(1,:)=A(1,:)+3 * A(3,:) \\
& A(2,:)=A(2,:)-3 * A(3,:)
\end{aligned}
$$

## Echelon matrices

## Example

Now, we'll repeat the same example using different row operations:

$$
\begin{aligned}
& \mathbf{r}_{1} \leftarrow \mathbf{r}_{3} \\
& \mathbf{r}_{1} \leftarrow-\mathbf{r}_{1} \\
& \mathbf{r}_{2} \leftarrow \mathbf{r}_{2}-4 \mathbf{r}_{1} \\
& \mathbf{r}_{3} \leftarrow \mathbf{r}_{3}-\mathbf{r}_{1} \\
& \hline
\end{aligned}\left|\left(\begin{array}{rrr}
1 & 2 & 3 \\
4 & 5 & 6 \\
-1 & -1 & 0 \\
-1 & -1 & 0 \\
4 & 5 & 6 \\
1 & 2 & 3
\end{array}\right) \quad \begin{array}{l}
\mathbf{r}_{1} \leftarrow \mathbf{r}_{1}-\mathbf{r}_{2} \\
\mathbf{r}_{3} \leftarrow \mathbf{r}_{3}-\mathbf{r}_{2} \\
1
\end{array}\right|\left(\begin{array}{lll}
\mathbf{r}_{3} \leftarrow-\frac{1}{3} \mathbf{r}_{3} \\
4 & 1 & 0 \\
1 & 2 & 6 \\
1 & 1 & 0 \\
0 & 1 & 6 \\
0 & 1 & 3
\end{array}\right) \quad\left(\begin{array}{rrr}
1 & 0 & -6 \\
0 & 1 & 6 \\
0 & 0 & -3 \\
1 & 0 & -6 \\
0 & 1 & 6 \\
0 & 0 & 1 \\
1 & 0 & 0 \\
0 & 1 & 0 \\
0 & 0 & 1
\end{array}\right)
$$

## Gauss-Jordan algorithm

## Definition 2.3 (Pivot and pivot column)

A pivot element is the element of a matrix that is used to perform certain calculations. For the Gauss-Jordan algorithm it corresponds to the first element different from zero in a given row. A pivot column is a column that contains a pivot.

## Step 1

Choose the left-most pivot column. The pivot element (marked in red) is any value within this column different from 0 . Note: Normally, we should take the one with maximum absolute value to avoid numerical errors.

## Example

$$
\left(\begin{array}{rrrrrr}
0 & 3 & -6 & 6 & 4 & -5 \\
3 & -7 & 8 & -5 & 8 & 9 \\
3 & -9 & 12 & -9 & 6 & 15
\end{array}\right)
$$

## Gauss-Jordan algorithm

## Step 2

Sort rows if necessary so that the pivot is as high as possible.

## Example

$$
\mathbf{r}_{3} \leftrightarrow \mathbf{r}_{1} \quad\left(\begin{array}{rrrrrr}
3 & -9 & 12 & -9 & 6 & 15 \\
3 & -7 & 8 & -5 & 8 & 9 \\
0 & 3 & -6 & 6 & 4 & -5
\end{array}\right)
$$

## Step 3

Use row operations to force the elements below the pivot to be 0 .

## Example

$$
\mathbf{r}_{2} \leftarrow \mathbf{r}_{2}-\mathbf{r}_{1} \left\lvert\,\left(\begin{array}{rrrrrr}
3 & -9 & 12 & -9 & 6 & 15 \\
0 & 2 & -4 & 4 & 2 & -6 \\
0 & 3 & -6 & 6 & 4 & -5
\end{array}\right)\right.
$$

## Gauss-Jordan algorithm

## Step 4

Repeat Steps 1 to 3 with the rows below the pivot.

## Example

$$
\mathbf{r}_{3} \leftarrow \mathbf{r}_{3}-\frac{2}{3} \mathbf{r}_{2}\left(\begin{array}{rrrrrr}
3 & -9 & 12 & -9 & 6 & 15 \\
0 & 3 & -6 & 6 & 4 & -5 \\
0 & 2 & -4 & 4 & 2 & -6
\end{array}\right)
$$

## Gauss-Jordan algorithm

## Step 5

Starting from the lowest and right-most pivot, force the elements above that pivot to be zero. If the pivot is not 1 , then rescale the row. Repeat with the next pivot on the left.

## Example

$$
\begin{array}{l|l}
\mathbf{r}_{3} \leftarrow-\frac{3}{2} r_{3} \\
r_{2} \leftarrow r_{2}-4 r_{3} \\
r_{1} \leftarrow r_{1}-6 r_{3} \\
r_{2} \leftarrow \frac{1}{3} r_{2} & \left(\begin{array}{rrrrrr}
3 & -9 & 12 & -9 & 6 & 15 \\
0 & 3 & -6 & 6 & 4 & -5 \\
0 & 0 & 0 & 0 & 1 & 4 \\
3 & -9 & 12 & -9 & 0 & -9 \\
0 & 3 & -6 & 6 & 0 & -21 \\
0 & 0 & 0 & 0 & 1 & 4 \\
3 & -9 & 12 & -9 & 0 & -9 \\
0 & 1 & -2 & 2 & 0 & -7 \\
0 & 0 & 0 & 0 & 0 & -7 \\
3 & 0 & -6 & 9 & 0 & -72 \\
0 & 1 & -2 & 2 & 0 & -7 \\
0 & 0 & -0 & 0 & 1 & -4 \\
1 & 0 & -r_{1}+9 r_{2} \\
1 & 0 & -2 & 3 & 0 & -24 \\
0 & 1 & -2 & 2 & 0 & -7 \\
0 & 0 & 0 & 0 & 1 & 4
\end{array}\right) \\
r_{1} \leftarrow \frac{1}{3} r_{1}
\end{array}
$$

Computing the inverse of a $n \times n$ matrix costs in the order of $n^{3}$ operations $\left(O\left(n^{3}\right)\right)$. However, calculating the reduced echelon form is only in the order of $n^{2}$ $\left(\mathrm{O}\left(n^{2}\right)\right)$. This difference is more and more important as $n$ grows.

## Existence and uniqueness of solutions (revisited)

We can now review the issue of existence and uniqueness under the light of the reduced echelon matrix.

## Example

$$
\left(\begin{array}{lll|l}
1 & 0 & 0 & 1 \\
0 & 1 & 0 & 4 \\
0 & 0 & 1 & 0
\end{array}\right)
$$

The system is compatible and the set of solutions is formed by a single point $S=\{(1,4,0)\}$.

## Example

$$
\left(\begin{array}{lll|l}
1 & 0 & 0 & 1 \\
0 & 1 & 1 & 4 \\
0 & 0 & 0 & 0
\end{array}\right)
$$

There are infinite solutions. The system is compatible indeterminate. The set of solutions is $S=\left\{\left(1,4-x_{3}, x_{3}\right) \forall x_{3} \in \mathbb{R}^{3}\right\}$. Because the set of solutions depends on a single variable, the set of solutions is a line.

## Existence and uniqueness of solutions (revisited)

## Example

$$
\left(\begin{array}{llll|l}
1 & 0 & 0 & 0 & 1 \\
0 & 1 & 1 & 1 & 4 \\
0 & 0 & 0 & 0 & 0
\end{array}\right)
$$

There are infinite solutions. The system is compatible indeterminate. The set of solutions is $S=\left\{\left(1,4-x_{3}-x_{4}, x_{3}, x_{4}\right) \forall x_{3}, x_{4} \in \mathbb{R}^{3}\right\}$. Now, the set of solutions depends on 2 variables and, consequently, it is a plane.

## Example

$$
\left(\begin{array}{lll|l}
1 & 0 & 0 & 1 \\
0 & 1 & 1 & 4 \\
0 & 0 & 0 & 1
\end{array}\right)
$$

The system is incompatible since the last equation is $0=1$. The set of solutions is the empty set, $S=\varnothing$.

## Exercises

Exercises<br>From Lay (4th ed.), Chapter 1, Section 2:<br>- 1.2.2<br>- 1.2.8<br>- 1.2.19<br>- 1.2.33<br>- 1.2.34

## Outline

(2) Linear equation system

- Introduction (a)
- Gauss-Jordan algorithm (b)
- Interpretation as a subspace (b)
- Existence and uniqueness of solutions (c)
- Applications (c)
- Linear independence (c)
- Linear transformations (d)
- Geometrical transformations (e)
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- More applications (e)


## Interpretation as a subspace

## Subspace spanned by columns

Consider the equation system given by the matricial equation $A \mathbf{x}=\mathbf{b}$, where $A \in \mathcal{M}_{n \times p}$. Let us call the $p$ columns of $A$ as $\mathbf{c}_{i} \in \mathbb{R}^{n}$. The previous equation can be rewritten as

$$
\left(\begin{array}{llll}
\mathbf{c}_{1} & \mathbf{c}_{2} & \ldots & \mathbf{c}_{p}
\end{array}\right)\left(\begin{array}{c}
x_{1} \\
x_{2} \\
\ldots \\
x_{p}
\end{array}\right)=\mathbf{b} \Rightarrow \sum_{i=1}^{p} x_{i} \mathbf{c}_{i}=\mathbf{b}
$$

That is, $A \mathbf{x}$ is the subspace spanned by the columns of matrix $A$.

$$
\operatorname{Span}\left\{\mathbf{c}_{1}, \mathbf{c}_{2}, \ldots, \mathbf{c}_{p}\right\}=\left\{\mathbf{v} \in \mathbb{R}^{n} \mid \mathbf{v}=A \mathbf{x} \forall \mathbf{x} \in \mathbb{R}^{p}\right\}
$$

The equation system $A \mathbf{x}=\mathbf{b}$ the poses the question: Find the weight coefficients $x_{i}$ such that vector $\mathbf{b}$ belongs to $\operatorname{Span}\left\{\mathbf{c}_{1}, \mathbf{c}_{2}, \ldots, \mathbf{c}_{p}\right\}$.

## Interpretation as a subspace

## Example

The equation system

$$
x_{1}+2 x_{2}-x_{3}=4
$$

can be represented as

$$
\left(\begin{array}{rrr}
1 & 2 & -1 \\
0 & -5 & 3
\end{array}\right)\left(\begin{array}{l}
x_{1} \\
x_{2} \\
x_{3}
\end{array}\right)=\binom{4}{1}
$$

That is, which are the weight coefficients $x_{1}, x_{2}$ and $x_{3}$ such that the vector $(4,1)$ belongs to the subspace generated by the vectors $(1,0),(2,-5)$, and $(-1,3)$.

## Interpretation as a subspace

## Theorem 3.1

The matrix equation $\mathbf{A x}=\mathbf{b}$ has the same solution as the vector equation $\sum_{i=1}^{p} x_{i} \mathbf{c}_{i}=\mathbf{b}$ and as the equation system whose augmented matrix is $\tilde{A}=(A \mid \mathbf{b})$.

## Theorem 3.2

For any $A \in \mathcal{M}_{n \times p}$ and vector $\mathbf{b} \in \mathbb{R}^{n}$, the following four statements are equivalent, that is, $P_{1} \Leftrightarrow P_{2} \Leftrightarrow P_{3} \Leftrightarrow P_{4}$
$P_{1}$ : The equation $A \mathbf{x}=\mathbf{b}$ has a solution.
$P_{2}$ : $\mathbf{b}$ is a linear combination of the columns of $A$.
$P_{3}$ : The columns of $A$ span all $\mathbb{R}^{n}$, i.e., $\operatorname{Span}\left\{\mathbf{c}_{i}\right\}=\mathbb{R}^{n}$.
$P_{4}$ : $A$ has a pivot in each row.

## Exercises

## Exercises

From Lay (4th ed.), Chapter 1, Section 4:

- 1.4.13
- 1.4.18
- 1.4.26
- 1.4.27
- 1.4.32
- 1.4.39
- 1.4.41 (bring computer)


## Outline

(2) Linear equation system

- Introduction (a)
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## Existence and uniqueness of solutions (revisited once again)

Let us consider the homogeneous system $A \mathbf{x}=\mathbf{0}$. It obviously has the trivial solution $\mathbf{x}=\mathbf{0}$. Non-trivial solutions can be found through the echelon matrix

## Example

$$
\begin{array}{rrr}
3 x_{1}+5 x_{2}-4 x_{3} & =0 \\
-3 x_{1} & -2 x_{2}+4 x_{3} & =0 \\
6 x_{1} & +x_{2} & -8 x_{3}
\end{array}=0 . \ldots \sim\left(\begin{array}{lll|l}
1 & 0 & \frac{4}{3} & 0 \\
0 & 1 & 0 & 0 \\
0 & 0 & 0 & 0
\end{array}\right)
$$

This is a compatible indeterminate system whose set of solutions is $S=\left\{\left(-\frac{4}{3} x_{3}, 0, x_{3}\right) \forall x_{3} \in \mathbb{R}\right\}$, or what is the same

$$
S=\operatorname{Span}\left\{\left(-\frac{4}{3}, 0,1\right)\right\} .
$$

That is, any of the infinite points in the straight line whose director vector is $\left(-\frac{4}{3}, 0,1\right)$ is a solution of the equation system.

## Existence and uniqueness of solutions (revisited once again)

Let us consider the non-homogeneous system $A \mathbf{x}=\mathbf{b}$.

## Example

$$
\begin{array}{rrrr}
3 x_{1}+5 x_{2} & -4 x_{3} & = & 7 \\
-3 x_{1} & -2 x_{2} & +4 x_{3} & = \\
6 x_{1} & +x_{2} & -8 x_{3} & = \\
-4
\end{array} \Rightarrow \ldots \sim\left(\begin{array}{lll|l}
1 & 0 & \frac{4}{3} & 0 \\
0 & 1 & 0 & 2 \\
0 & 0 & 0 & 0
\end{array}\right)
$$

This is a compatible indeterminate system whose set of solutions is $S=\left\{\left(-\frac{4}{3} x_{3}, 2, x_{3}\right) \forall x_{3} \in \mathbb{R}\right\}$, or what is the same

$$
S=\left\{(0,2,0)+\left(-\frac{4}{3} x_{3}, 0, x_{3}\right) \forall x_{3} \in \mathbb{R}\right\}=(0,2,0)+\operatorname{Span}\left\{\left(-\frac{4}{3}, 0,1\right)\right\} .
$$

That is, any of the infinite points in the straight line whose director vector is $\left(-\frac{4}{3}, 0,1\right)$ and passes through the point $(0,2,0)$ is a solution of the equation system.

## Existence and uniqueness of solutions (revisited once again)

Consider the following homogeneous equation system

## Example

$$
10 x_{1}-3 x_{2}-2 x_{3}=0 \Rightarrow \ldots \sim\left(\begin{array}{ccc}
10 & -3 & -2 \mid 0
\end{array}\right)
$$

This is a compatible indeterminate system whose set of solutions is $S=\left\{\left(\frac{3}{10} x_{2}+\frac{1}{5} x_{3}, x_{2}, x_{3}\right) \forall x_{2}, x_{3} \in \mathbb{R}\right\}$, or what is the same

$$
S=\operatorname{Span}\left\{\left(\frac{3}{10}, 1,0\right),\left(\frac{1}{5}, 0,1\right)\right\} .
$$

That is, any of the infinite points in the plane containing the vectors ( $\frac{3}{10}, 1,0$ ) and $\left(\frac{1}{5}, 0,1\right)$ is a solution of the equation system.

## Existence and uniqueness of solutions (revisited once again)

Consider now the following non-homogeneous equation system

## Example

$$
10 x_{1}-3 x_{2}-2 x_{3}=10 \Rightarrow \ldots \sim\left(\begin{array}{ccc|c}
10 & -3 & -2 \mid 10
\end{array}\right)
$$

This is a compatible indeterminate system whose set of solutions is $S=\left\{\left(1+\frac{3}{10} x_{2}+\frac{1}{5} x_{3}, x_{2}, x_{3}\right) \forall x_{2}, x_{3} \in \mathbb{R}\right\}$, or what is the same

$$
\begin{gathered}
S=\left\{(1,0,0)+\left(\frac{3}{10} x_{2}+\frac{1}{5} x_{3}, x_{2}, x_{3}\right) \forall x_{2}, x_{3} \in \mathbb{R}\right\}= \\
(1,0,0)+\operatorname{Span}\left\{\left(\frac{3}{10}, 1,0\right),\left(\frac{1}{5}, 0,1\right)\right\} .
\end{gathered}
$$

## Existence and uniqueness of solutions (revisited once again)

## Corollary 4.1

Consider the compatible, non-homogeneous equation system given by $A \mathbf{x}=\mathbf{b}$ and its homogeneous counterpart $A \mathbf{x}=\mathbf{0}$. Let $S_{h}$ be the set of solutions of the homogeneous equation system. Then, the set of solutions of the non-homogeneous equation system is of the form

$$
S_{n h}=\mathbf{x}_{0}+S_{h}
$$

For some $\mathbf{x}_{0} \in \mathbb{R}^{n}$.

## Definition 4.1 (Null space of $A$ )

$S_{h}$ is called the null space of the matrix $A$. It has the property that given an equation system $A \mathbf{x}=\mathbf{b}$, if $\mathbf{x}_{0}$ is a solution of the equation system, then $\mathbf{x}_{0}+\mathbf{x}_{h}$ is also a solution, for any $\mathbf{x}_{h} \in S_{h}$.

## Existence and uniqueness of solutions (revisited once again)


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Uppsala University, Department of Neuroscience

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## Existence and uniqueness of solutions (revisited once again)

In this example, the authors describe how to solve a problem appearing in the tomographic use of a certain microscope due to the absence of some measurements (resulting in an important null space of the tomographic problem).



## Existence and uniqueness of solutions (revisited once again)

In this example, the authors describe how the exact location of a tooth fracture is uncertain (Fig. C) due to the artifacts introduced by the null space of the tomographic problem.


Mora, M. A.; Mol, A.; Tyndall, D. A., Rivera, E. M. In vitro assessment of local computed tomography for the detection of longitudinal tooth fractures.
Oral Surg Oral Med Oral Pathol Oral Radiol Endod, 2007, 103, 825-829.

## Exercises

## Exercises

From Lay (4th ed.), Chapter 1, Section 5:

- 1.5.11
- 1.5.13
- 1.5.19
- 1.5.21
- 1.5.25
- 1.5.26
- 1.5.36
- 1.5.39


## Outline

(2) Linear equation system

- Introduction (a)
- Gauss-Jordan algorithm (b)
- Interpretation as a subspace (b)
- Existence and uniqueness of solutions (c)
- Applications (c)
- Linear independence (c)
- Linear transformations (d)
- Geometrical transformations (e)
- Classification of functions (e)
- More applications (e)


## Applications

In fluorescence microscopy, we can quantitatively measure the amount of fluorescence coming from each source with a linear equation system.


$$
\left(\begin{array}{l}
B g \\
B \\
T \\
S
\end{array}\right)=\left(\begin{array}{llll}
1 & 0 & 0 & 0 \\
1 & \frac{1}{2} & 0 & 0 \\
1 & 0 & \frac{1}{2} & 0 \\
1 & \frac{1}{2} & 0 & \frac{1}{2}
\end{array}\right)\left(\begin{array}{l}
\widehat{B g} \\
\hat{B} \\
\hat{T} \\
\hat{T_{s}}
\end{array}\right)
$$

C. Calabia-Linares, M. Pérez-Martínez, N. Martín-Cofreces, M. Alfonso-Pérez, C. Gutiérrez-Vázquez, M. Mittelbrunn, S. Ibiza, F.R. Urbano-Olmos, C. Aguado-Ballano, C.O.S. Sorzano, F. Sánchez-Madrid, E. Veiga. Clathrin drives actin accumulation at the immunological synapse. J. Cell Science, 124: 820-830 (2011)

## Applications

In computed tomography, a simple model (but widely used) for data collection states that the data observed is the sum of the values of the density found along the X -ray path.

## FIGURE 25-14

CT data acquisition. A simple CT system passes a narrow beam of x-rays through the body from source to detector. The source and detector are then translated to obtain a complete view. The remaining views are obtained by rotating the source and detector in about $1^{\circ}$ increments, and repeating the translation process.


## Applications

In the blood system, at each node, the sum of output flows must be equal to the sum of input flows.


## Applications

In a very simplified model, respiration is the burning of glucose that can be written as

$$
x_{1} \mathrm{C}_{6} \mathrm{H}_{12} \mathrm{O}_{6}+x_{2} \mathrm{O}_{2} \rightarrow x_{3} \mathrm{CO}_{2}+x_{4} \mathrm{H}_{2} \mathrm{O}
$$



C: $6 x_{1}=x_{3}$
H: $6 x_{1}=2 x_{4}$
O: $6 x_{1}+2 x_{2}=2 x_{3}+x_{4}$

## Exercises

## Exercises

From Lay (4th ed.), Chapter 1, Section 6:

- 1.6.5
- 1.6.7
- 1.6.12


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## Linear independence

## Definition 6.1 (Linear independence)

A set of vectors $\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{p}$ is linearly independent if

$$
x_{1} \mathbf{v}_{1}+x_{2} \mathbf{v}_{2}+\ldots+x_{p} \mathbf{v}_{p}=\mathbf{0} \Rightarrow x_{1}=x_{2}=\ldots=x_{p}=0
$$

That is the only solution of the previous equation is the trivial solution $\mathbf{x}=\mathbf{0}$. The set is linearly dependent if at least two $x_{i}$ 's are different from 0 .

## Linear independence

## Example

Determine if the vectors $\mathbf{v}_{1}=(1,2,3), \mathbf{v}_{2}=(4,5,6)$, and $\mathbf{v}_{3}=(2,1,0)$ are linearly independent.

## Solution

The augmented matrix associated to the equation system in Definition 6.1 is

$$
\left(\begin{array}{lll|l}
1 & 4 & 2 & 0 \\
2 & 5 & 1 & 0 \\
3 & 6 & 0 & 0
\end{array}\right) \sim \ldots \sim\left(\begin{array}{rrr|r}
1 & 4 & 2 & 0 \\
0 & -3 & -3 & 0 \\
0 & 0 & 0 & 0
\end{array}\right)
$$

Since the system is compatible indeterminate, there exists a solution apart from the trivial solution and, therefore, the vectors are linearly dependent.

## Linear independence

## Example

If possible, find a linear relationship among the three vectors. Solution We continue transforming the augmented matrix to its reduced echelon form

$$
\left(\begin{array}{rrr|r}
1 & 4 & 2 & 0 \\
0 & -3 & -3 & 0 \\
0 & 0 & 0 & 0
\end{array}\right) \sim \ldots \sim\left(\begin{array}{rrr|r}
1 & 0 & -2 & 0 \\
0 & 1 & 1 & 0 \\
0 & 0 & 0 & 0
\end{array}\right)
$$

From which $x_{1}=2 x_{3}$ and $x_{2}=-x_{3}$. Simply by choosing $x_{3}=1$, we obtain have that a possible solution to the equation system in Definition 6.1 is $x_{1}=2$,
$x_{2}=-1$ and $x_{3}=1$, consequently we have that

$$
2 \mathbf{v}_{1}-\mathbf{v}_{2}+\mathbf{v}_{3}=\mathbf{0}
$$

## Linear independence

## Example

$\mathbf{v}_{1}=(3,1)$ and $\mathbf{v}_{2}=(6,2)$ are linearly dependent because

$$
\mathbf{v}_{2}=2 \mathbf{v}_{1} \Rightarrow-2 \mathbf{v}_{1}+\mathbf{v}_{2}=\mathbf{0} \Rightarrow \mathbf{v}_{1}=\frac{1}{2} \mathbf{v}_{2}
$$



If two vectors are linearly dependent of each other, then any one of them is a multiple of the other.

## Linear independence

## Example

$\mathbf{v}_{1}=(3,2)$ and $\mathbf{v}_{2}=(6,2)$ are linearly independent


## Linear independence

## Example

$\mathbf{v}_{1}=(1,1,0), \mathbf{v}_{2}=(-1,1,0)$ and $\mathbf{v}_{3}=(0,2,0)$ are linearly dependent because

$$
\mathbf{v}_{3}=\mathbf{v}_{1}+\mathbf{v}_{2}
$$



## Linear independence

## Example

$$
\mathbf{v}_{1}=(1,1,0), \mathbf{v}_{2}=(-1,1,0) \text { and } \mathbf{v}_{3}=(0,2,1) \text { are linearly independent }
$$



## Linear independence

## Theorem 6.1 (Linear independence of matrix columns)

The columns of the matrix $A$ are linearly independent iff the only solution of $A \mathbf{x}=\mathbf{0}$ is the trivial one.
Proof
Let $A=\left[\begin{array}{lll}\mathbf{a}_{1} & \mathbf{a}_{2} & \ldots\end{array} \mathbf{a}_{p}\right]$ so that the columns of the matrix $A$ are the vectors $\mathbf{a}_{i}$. According to Definition 6.1 these vectors are linearly independent iff

$$
x_{1} \mathbf{a}_{1}+x_{2} \mathbf{a}_{2}+\ldots+x_{p} \mathbf{a}_{p}=\mathbf{0} \Rightarrow x_{1}=x_{2}=\ldots=x_{p}=0
$$

or what is the same

$$
A \mathbf{x}=\mathbf{0} \Rightarrow x_{1}=x_{2}=\ldots=x_{p}=0
$$

as stated by the theorem (q.e.d.)

## Linear independence

## Theorem 6.2

Any set $\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{p}\right\}$ with $\mathbf{v}_{i} \in \mathbb{R}^{n}$ is linearly dependent if $p>n$. Proof
Let $A=\left[\mathbf{v}_{1} \mathbf{v}_{2} \ldots \mathbf{v}_{p}\right]$ and let us consider the equation system $A \mathbf{x}=\mathbf{0}$. If $p>n$ there are more unknowns than equations, and consequently, there are free variables and the system is compatible indeterminate. Thus, there are more solutions apart from the trivial one and the set of vectors is linearly dependent.

## Theorem 6.3

If any set $\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{p}\right\}$ with $\mathbf{v}_{i} \in \mathbb{R}^{n}$ contains the vector $\mathbf{0}$, then the set of vectors is linearly dependent.
Proof
We can assume, without loss of generality, that $\mathbf{v}_{1}=\mathbf{0}$. Then, we can set $x_{1}=1$, $x_{2}=x_{3}=\ldots=x_{p}=0$ so that the following equation is met:

$$
1 \mathbf{v}_{1}+0 \mathbf{v}_{2}+\ldots+0 \mathbf{v}_{p}=\mathbf{0} \text { (q.e.d.) }
$$

## Linear independence

## Theorem 6.4

A set of vectors is linearly dependent iff at least 1 of the vectors is linearly dependent on the rest
Proof
Proof $\Leftarrow$
Let us assume that $\mathbf{v}_{j}$ is a linear combination of the rest of the vectors, that is,

$$
\mathbf{v}_{j}=\sum_{k \neq j} x_{k} \mathbf{v}_{k}
$$

Then, we can write $\mathbf{v}_{j}-\sum_{k \neq j} x_{k} \mathbf{v}_{k}=\mathbf{0} \Rightarrow$

$$
-x_{1} \mathbf{v}_{1}-x_{2} \mathbf{v}_{2}-\ldots-x_{j-1} \mathbf{v}_{j-1}+\mathbf{v}_{j}-x_{j+1} \mathbf{v}_{j+1}-x_{p} \mathbf{v}_{p}=\mathbf{0}
$$

And consequently there exists a non-trivial solution of the equation of Definition 6.1.

## Linear independence

## Proof $\Rightarrow$

If $\mathbf{v}_{1}=\mathbf{0}$, then we have already a vector that is a trivial combination of the rest $\left(\mathbf{v}_{1}=0 \mathbf{v}_{2}+0 \mathbf{v}_{3}+\ldots+0 \mathbf{v}_{p}\right)$.
If $\mathbf{v}_{1} \neq \mathbf{0}$, then there exist some coefficients such that

$$
x_{1} \mathbf{v}_{1}+x_{2} \mathbf{v}_{2}+\ldots+x_{p} \mathbf{v}_{p}=\mathbf{0}
$$

Let $j$ be the largest index for which $x_{j} \neq 0$ (that is, $x_{j+1}=x_{j+2}=\ldots=$ $x_{p}=0$ ).
If $j=1$, then $x_{1} \mathbf{v}_{1}=\mathbf{0}$, but this is not possible because $\mathbf{v}_{1} \neq \mathbf{0}$. Then, $j>1$ and consequently

$$
\begin{gathered}
x_{1} \mathbf{v}_{1}+x_{2} \mathbf{v}_{2}+\ldots+x_{j} \mathbf{v}_{j}+0 \mathbf{v}_{j+1}+\ldots+0 \mathbf{v}_{p}=\mathbf{0} \Rightarrow \\
x_{j} \mathbf{v}_{j}=-x_{1} \mathbf{v}_{1}-x_{2} \mathbf{v}_{2}-\ldots-x_{j-1} \mathbf{v}_{j-1} \Rightarrow \\
\mathbf{v}_{j}=-\frac{x_{1}}{x_{j}} \mathbf{v}_{1}-\frac{x_{2}}{x_{j}} \mathbf{v}_{2}-\ldots-\frac{x_{j-1}}{x_{j}} \mathbf{v}_{j-1} \text { (q.e.d.) }
\end{gathered}
$$

## Exercises

## Exercises

From Lay (4th ed.), Chapter 1, Section 7:

- 1.7.9
- 1.7.39
- 1.7.40
- 1.7.41 (bring computer)


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## Linear transformations

## Definition 7.1 (Transformation)

A transformation (or function or mapping), $T$, from $\mathbb{R}^{n}$ to $\mathbb{R}^{m}$ is a rule that assigns to each vector of $\mathbb{R}^{n}$ a vector of $\mathbb{R}^{m}$.

$$
\begin{aligned}
T: \mathbb{R}^{n} & \rightarrow \mathbb{R}^{m} \\
\mathbf{x} & \rightarrow T(\mathbf{x})
\end{aligned}
$$

$\mathbb{R}^{n}$ is called the domain of the transformation, and $\mathbb{R}^{m}$ its codomain. $T(\mathbf{x})$ is the image of vector $\mathbf{x}$ under the action of $T$. The set of all images is the range of $T$.


## Linear transformations

## Definition 7.2 (Matrix transformation)

$T$ is a matrix transformation iff $T(\mathbf{x})=A \mathbf{x}$ for some matrix $A \in \mathcal{M}_{m \times n}$.

## Example

Let us consider $A=\left(\begin{array}{cccc}4 & -3 & 1 & 3 \\ 2 & 0 & 5 & 1\end{array}\right)$ and the matrix transformation $\mathbf{y}=A \mathbf{x}$. For instance, the image of $\mathbf{x}=(1,1,1,1)$ is

$$
\mathbf{y}=\left(\begin{array}{cccc}
4 & -3 & 1 & 3 \\
2 & 0 & 5 & 1
\end{array}\right)\left(\begin{array}{l}
1 \\
1 \\
1 \\
1
\end{array}\right)=\binom{5}{8}
$$

The equation system $A \mathbf{x}=\binom{5}{8}$ looks for all those $\mathbf{x}$, if any, such that $T(\mathbf{x})=\binom{5}{8}$. The domain of this transformation is $\mathbb{R}^{4}$ and its codomain $\mathbb{R}^{2}$.

## Linear transformations

## Example

Let us consider $A=\left(\begin{array}{ll}1 & 0 \\ 0 & 1 \\ 0 & 0\end{array}\right)$ and the matrix transformation $\mathbf{y}=A \mathbf{x}$. The domain of this transformation is $\mathbb{R}^{2}$ and its codomain $\mathbb{R}^{3}$. However, not all points in $\mathbb{R}^{3}$ need to be an image of some point $\mathbf{x} \in \mathbb{R}^{2}$, only a subset of them may be. In this case,

$$
\mathbb{R}^{3} \supset \operatorname{Range}(T)=\langle(1,0,0),(0,1,0)\rangle
$$

In general, the range of the transformation $T$ is the subspace spanned by the columns of the matrix $A$.

## Linear transformations

## Example

Let us consider $A=\left(\begin{array}{cc}1 & -3 \\ 3 & 5 \\ -1 & 7\end{array}\right)$ and the matrix transformation $\mathbf{y}=A \mathbf{x}$.
(1) What is the image of $\mathbf{u}=(2,-1)$ under $T$ ?

$$
T(\mathbf{u})=A \mathbf{u}=(5,1,9)
$$

(2) Let $\mathbf{b}=(3,2,-5)$. Which is $\mathbf{x}$ such that $T(\mathbf{x})=\mathbf{b}$ ?

$$
\left(\begin{array}{rr|r}
1 & -3 & 3 \\
3 & 5 & 2 \\
-1 & 7 & -5
\end{array}\right) \sim\left(\begin{array}{rr|r}
1 & 0 & \frac{3}{2} \\
0 & 1 & -\frac{1}{2} \\
0 & 0 & 0
\end{array}\right)
$$

From which we deduce $\mathbf{x}=\left(\frac{3}{2},-\frac{1}{2}\right)$.
(3) Is there any other $\mathbf{x}$ such that $T(\mathbf{x})=\mathbf{b}$ ?

No, the previous solution is unique because the equation system is definite compatible.

## Linear transformations

## Example

( Does $\mathbf{c}=(3,2,5)$ belong to Range $(T)$ ?
$\left(\begin{array}{rr|r}1 & -3 & 3 \\ 3 & 5 & 2 \\ -1 & 7 & 5\end{array}\right) \sim\left(\begin{array}{rr|r}1 & 0 & \frac{3}{2} \\ 0 & 1 & -\frac{1}{2} \\ 0 & 0 & -35\end{array}\right)$

Since the system is incompatible, we deduce that there is no vector $\mathbf{x}$ whose image is $\mathbf{c}$ and, consequently, $\mathbf{c} \notin \operatorname{Range}(T)$.
(0) Which is the function $\mathbf{y}=T(\mathbf{x})$ ?

$$
\left(\begin{array}{l}
y_{1} \\
y_{2} \\
y_{3}
\end{array}\right)=\left(\begin{array}{cc}
1 & -3 \\
3 & 5 \\
-1 & 7
\end{array}\right)\binom{x_{1}}{x_{2}}=\left(\begin{array}{c}
x_{1}-3 x_{2} \\
3 x_{1}+5 x_{2} \\
-x_{1}+7 x_{2}
\end{array}\right)
$$

(0. Which is Range $(T)$ ?

Range $(T)=\langle(1,3,-1),(-3,5,7)\rangle=$
$\left\{\mathbf{y} \in \mathbb{R}^{3} \left\lvert\, \mathbf{y}=x_{1}\left(\begin{array}{c}1 \\ 3 \\ -1\end{array}\right)+x_{2}\left(\begin{array}{c}-3 \\ 5 \\ 7\end{array}\right) \forall x_{1}\right., x_{2} \in \mathbb{R}\right\}$
Because $(1,3,-1)$ and $(-3,5,7)$ are linearly independent, $\operatorname{Range}(T)$ is a plane.

## Linear transformations

## Example

Consider the transformation $T(\mathbf{x})=\left(\begin{array}{lll}1 & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 0 & 0\end{array}\right)\left(\begin{array}{l}x_{1} \\ x_{2} \\ x_{3}\end{array}\right)=\left(\begin{array}{c}x_{1} \\ x_{2} \\ 0\end{array}\right)$. This is a projection transformation that projects any 3D point onto the $X Y$ plane.


## Linear transformations

## Definition 7.3 (Linear transformation)

$T$ is a linear transformation iff $\forall \mathbf{x}_{1}, \mathbf{x}_{2} \in \operatorname{Dom}(T), \forall c \in \mathbb{R}$
(1) $T\left(\mathbf{x}_{1}+\mathbf{x}_{2}\right)=T\left(\mathbf{x}_{1}\right)+T\left(\mathbf{x}_{2}\right)$
(2) $T\left(c \mathbf{x}_{1}\right)=c T\left(\mathbf{x}_{1}\right)$

## Theorem 7.1

If $T(\mathbf{x})$ is a linear transformation, then
(1) $T(\mathbf{0})=\mathbf{0}$
(2) $T\left(c_{1} \mathbf{x}_{1}+c_{2} \mathbf{x}_{2}\right)=c_{1} T\left(\mathbf{x}_{1}\right)+c_{2} T\left(\mathbf{x}_{2}\right) \forall \mathbf{x}_{1}, \mathbf{x}_{2} \in \operatorname{Dom}(T), \forall c_{1}, c_{2} \in \mathbb{R}$

Proof
(1) $T(\mathbf{0})=T\left(0 \mathbf{x}_{1}\right)=[(2)$, Def. 7.3 $]=0 T\left(\mathbf{x}_{1}\right)=\mathbf{0}$ (q.e.d.)
(2) $T\left(c_{1} \mathbf{x}_{1}+c_{2} \mathbf{x}_{2}\right)=\left[(1)\right.$, Def. 7.3] $=T\left(c_{1} \mathbf{x}_{1}\right)+T\left(c_{2} \mathbf{x}_{2}\right)=[(2)$, Def. 7.3] $c_{1} T\left(\mathbf{x}_{1}\right)+c_{2} T\left(\mathbf{x}_{2}\right)$ (q.e.d.)

## Linear transformations

## Theorem 7.2

If $\forall \mathbf{x}_{1}, \mathbf{x}_{2} \in \operatorname{Dom}(T), \forall c_{1}, c_{2} \in \mathbb{R}$ it is verified that
$T\left(c_{1} \mathbf{x}_{1}+c_{2} \mathbf{x}_{2}\right)=c_{1} T\left(\mathbf{x}_{1}\right)+c_{2} T\left(\mathbf{x}_{2}\right)$, then $T(\mathbf{x})$ is a linear transformation.
Proof
(1) Let us consider the case $c_{1}=c_{2}=1$, then according to the assumption of the theorem we have $T\left(\mathbf{x}_{1}+\mathbf{x}_{2}\right)=T\left(\mathbf{x}_{1}\right)+T\left(\mathbf{x}_{2}\right)$, which implies (1) in Def. 7.3.
(2) Let us consider the case $c_{2}=0$, then according to the assumption of the theorem we have $T\left(c_{1} \mathbf{x}_{1}\right)=c_{1} T\left(\mathbf{x}_{1}\right)$, which implies (2) in Def. 7.3.
(q.e.d.)

## Linear transformations

## Corollary: Principle of superposition

If $\forall \mathbf{x}_{i} \in \operatorname{Dom}(T), \forall c_{i} \in \mathbb{R}$ it is verified that $T\left(\sum_{i} c_{i} \mathbf{x}_{i}\right)=\sum_{i} c_{i} T\left(\mathbf{x}_{i}\right)$.
Proof
Apply the previous theorem multiple times. (q.e.d.)

## Linear transformations

## Example

Show that $T(\mathbf{x})=\left(\begin{array}{cc}1 & 0 \\ 0 & -1\end{array}\right)\binom{x_{1}}{x_{2}}$ is a linear transformation.

## Proof

(1) Show that $T\left(\mathbf{x}_{1}+\mathbf{x}_{2}\right)=T\left(\mathbf{x}_{1}\right)+T\left(\mathbf{x}_{2}\right)$

On one side we have $T\left(\mathbf{x}_{1}+\mathbf{x}_{2}\right)=\left(\begin{array}{cc}1 & 0 \\ 0 & -1\end{array}\right)\binom{x_{11}+x_{21}}{x_{12}+x_{22}}=\binom{x_{11}+x_{21}}{-x_{12}-x_{22}}$
On the other side we have $T\left(\mathbf{x}_{1}\right)+T\left(\mathbf{x}_{2}\right)=$

$$
\left(\begin{array}{cc}
1 & 0 \\
0 & -1
\end{array}\right)\binom{x_{11}}{x_{12}}+\left(\begin{array}{cc}
1 & 0 \\
0 & -1
\end{array}\right)\binom{x_{21}}{x_{22}}=\binom{x_{11}}{-x_{12}}+\binom{x_{21}}{-x_{22}}=\binom{x_{11}+x_{21}}{-x_{12}-x_{22}}
$$

Obviously, these two calculations give the same result.
(2) Show that $T\left(c_{1} \mathbf{x}_{1}\right)=c_{1} T\left(\mathbf{x}_{1}\right)$

$$
\begin{aligned}
T\left(c_{1} \mathbf{x}_{1}\right) & =\left(\begin{array}{cc}
1 & 0 \\
0 & -1
\end{array}\right)\binom{c_{1} x_{11}}{c_{1} x_{12}}=\binom{c_{1} x_{11}}{-c_{1} x_{12}}=c_{1}\binom{x_{11}}{-x_{12}} \\
& =c_{1}\left(\begin{array}{cc}
1 & 0 \\
0 & -1
\end{array}\right)\binom{x_{11}}{x_{12}}=c_{1} T\left(\mathbf{x}_{1}\right)
\end{aligned}
$$

## Linear transformations

## Theorem 7.3

Any matrix transformation is a linear transformation.
Proof
(1) Show that $T\left(\mathbf{x}_{1}+\mathbf{x}_{2}\right)=T\left(\mathbf{x}_{1}\right)+T\left(\mathbf{x}_{2}\right)$

$$
T\left(\mathbf{x}_{1}+\mathbf{x}_{2}\right)=A\left(\mathbf{x}_{1}+\mathbf{x}_{2}\right)=A \mathbf{x}_{1}+A \mathbf{x}_{2}=T\left(\mathbf{x}_{1}\right)+T\left(\mathbf{x}_{2}\right) \text { (q.e.d.) }
$$

(2) Show that $T\left(c_{1} \mathbf{x}_{1}\right)=c_{1} T\left(\mathbf{x}_{1}\right)$

$$
T\left(c_{1} \mathbf{x}_{1}\right)=A\left(c_{1} \mathbf{x}_{1}\right)=c_{1}\left(A \mathbf{x}_{1}\right)=c_{1} T\left(\mathbf{x}_{1}\right) \text { (q.e.d.) }
$$

## Reinterpreting the columns of a matrix

## Example

Consider $T(\mathbf{x})=A \mathbf{x}$ with $A=\left(\begin{array}{cccc}4 & -3 & 1 & 3 \\ 2 & 0 & 5 & 1\end{array}\right)$. Consider the standard canonical basis of $\mathbb{R}^{4}$ formed by the vectors $\mathbf{e}_{1}=(1,0,0,0), \mathbf{e}_{2}=(0,1,0,0)$, $\mathbf{e}_{3}=(0,0,1,0)$, and $\mathbf{e}_{4}=(0,0,0,1)$. Let us consider the transformation of each one of these vectors

$$
T\left(\mathbf{e}_{1}\right)=\binom{4}{2} \quad T\left(\mathbf{e}_{2}\right)=\binom{-3}{0} \quad T\left(\mathbf{e}_{3}\right)=\binom{1}{5} \quad T\left(\mathbf{e}_{4}\right)=\binom{3}{1}
$$

In general, we note that the transformation of $\mathbf{e}_{i}$ is the $i$-th column of matrix $A$.

## Corollary

The columns of the matrix $A \in \mathcal{M}_{m \times n}$ can be understood as the transformations of the canonical basis of $\mathbb{R}^{n}$ :

$$
A=\left(\begin{array}{llll}
\mathbf{a}_{1} & \mathbf{a}_{2} & \ldots & \mathbf{a}_{n}
\end{array}\right)=\left(\begin{array}{llll}
T\left(\mathbf{e}_{1}\right) & T\left(\mathbf{e}_{2}\right) & \ldots & T\left(\mathbf{e}_{n}\right)
\end{array}\right)
$$

## Reinterpreting the columns of a matrix

## Example (continued)

In the previous example consider transforming the vector $\mathbf{x}=(1,-2,3,5)$. This vector is equal to

$$
\mathbf{x}=\mathbf{e}_{1}-2 \mathbf{e}_{2}+3 \mathbf{e}_{3}+5 \mathbf{e}_{4}
$$

Then, we have

$$
\begin{aligned}
T(\mathbf{x}) & =T\left(\mathbf{e}_{1}-2 \mathbf{e}_{2}+3 \mathbf{e}_{3}+5 \mathbf{e}_{4}\right)=T\left(\mathbf{e}_{1}\right)-2 T\left(\mathbf{e}_{2}\right)+3 T\left(\mathbf{e}_{3}\right)+5 T\left(\mathbf{e}_{4}\right) \\
& =\binom{4}{2}-2\binom{-3}{0}+3\binom{1}{5}+5\binom{3}{1}=\binom{28}{22}
\end{aligned}
$$

## Exercises

Exercises<br>From Lay (4th ed.), Chapter 1, Section 8:<br>- 1.8.23<br>- 1.8.25<br>- 1.8.26<br>- 1.8.30<br>- 1.8.34

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## Geometrical transformations

Certain matrix transformations are used to transform the unit square into different shapes. The following table shows some of such transformations.


Transformation
Reflection through the $x_{1}$-axis

Image of the Unit Square


Standard Matrix
$\left[\begin{array}{rr}1 & 0 \\ 0 & -1\end{array}\right]$

## Geometrical transformations

Reflection through the $x_{2}$-axis


$\left[\begin{array}{rr}-1 & 0 \\ 0 & 1\end{array}\right]$

Reflection through
the line $x_{2}=x_{1}$
$\left[\begin{array}{ll}0 & 1 \\ 1 & 0\end{array}\right]$


## Geometrical transformations

Reflection through
the line $x_{2}=-x_{1}$


$$
\left[\begin{array}{rr}
0 & -1 \\
-1 & 0
\end{array}\right]
$$

Reflection through the origin


$$
\left[\begin{array}{rr}
-1 & 0 \\
0 & -1
\end{array}\right]
$$

## Geometrical transformations

TABLE 2 Contractions and Expansions

| Transformation |  |
| :--- | :--- |
| Horizontal <br> contraction <br> and expansion | Image of the Unit Square |$\quad$ Standard Matri



## Geometrical transformations

TABLE 3 Shears


Vertical shear



$$
\left[\begin{array}{ll}
1 & 0 \\
k & 1
\end{array}\right]
$$

$$
k<0
$$

$k>0$

## Geometrical transformations

TABLE 4 Projections

| Transformation | Image of the Unit Square | Standard Matrix |
| :--- | :---: | :---: |
| Projection onto <br> the $x_{1}$-axis |  | $\left[\begin{array}{ll}1 & 0 \\ 0 & 0\end{array}\right]$ |

Projection onto the $x_{2}$-axis


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- Existence and uniqueness of solutions (c)
- Applications (c)
- Linear independence (c)
- Linear transformations (d)
- Geometrical transformations (e)
- Classification of functions (e)
- More applications (e)


## Classification of functions

## Definition 9.1

Functions can be classified as surjective, injective or bijective:
Surjective: A function is surjective if every point of the codomain has at least one point of the domain that maps onto it. They are also called onto functions.

Injective: A function is injective if every point of the codomain has at most one point in the domain that maps onto it. They are also called one-to-one functions.
Bijective: A function is bijective if it is injective and surjective.


## Classification of functions

## Example

Here we have some examples of the classification of functions applied to linear transformations

$T$ is not onto $\mathbb{R}^{m}$

$\qquad$

$T$ is onto $\mathbb{R}^{m}$

FIGURE 3 Is the range of $T$ all of $\mathbb{R}^{m}$ ?

$T$ is not one-to-one
$T$ is one-to-one
FIGURE 4 Is every b the image of at most one vector?

## Classification of functions

## Example

Consider $T(\mathbf{x})=A \mathbf{x}$ with $A=\left(\begin{array}{cccc}1 & -4 & 8 & 1 \\ 0 & 2 & -1 & 3 \\ 0 & 0 & 0 & 5\end{array}\right)$. This is a transformation from $\mathbb{R}^{4}$ onto $\mathbb{R}^{3}$. The columns of $A \mathbf{a}_{1}, \mathbf{a}_{2}$, and $\mathbf{a}_{4}$ are linearly independent and span $\mathbb{R}^{3}$ (that is, the function is surjective). Therefore, there must be points in $\mathbb{R}^{3}$ that come from several points in $\mathbb{R}^{4}$ (the function is not injective). Let us find some of these points.

$$
\begin{gathered}
\left(\begin{array}{rrrr|r}
1 & -4 & 8 & 1 & y_{1} \\
0 & 2 & -1 & 3 & y_{2} \\
0 & 0 & 0 & 5 & y_{3}
\end{array}\right) \sim\left(\begin{array}{rrrr|c}
1 & 0 & 6 & 0 & y_{1}-2 y_{2}-\frac{4}{5} y_{3} \\
0 & 1 & -\frac{1}{2} & 0 & \frac{1}{2} y_{2}-\frac{3}{10} y_{3} \\
0 & 0 & 0 & 1 & \frac{1}{5} y_{3}
\end{array}\right) \Rightarrow \\
x_{1}=y_{1}-2 y_{2}-\frac{4}{5} y_{3}-6 x_{3} \\
x_{2}=\frac{1}{2} y_{2}-\frac{3}{10} y_{3}+\frac{1}{2} x_{3} \\
x_{4}=\frac{1}{5} y_{3}
\end{gathered}
$$

Since $x_{3}$ is a free variable, we have that for each point in the codomain, there is a straight line that maps onto it (the equation of the line is the one given above).

## Classification of functions

## Theorem 9.1

Let $T(\mathbf{x})$ be a linear transformation. $T(\mathbf{x})$ is an injective function iff $T(\mathbf{x})=\mathbf{0}$ has only the trivial solution $\mathbf{x}=\mathbf{0}$.

## Proof

Proof $\Rightarrow$
If $T$ is injective, then, by definition, every point of the codomain, in particular $\mathbf{0}$ is the mapping of at most one point in the domain. We already know that for any linear transformation $T(\mathbf{0})=\mathbf{0}$, therefore, $\mathbf{x}=\mathbf{0}$ must be the unique solution of the equation $T(\mathbf{x})=\mathbf{0}$.

## Classification of functions

Proof $\Leftarrow$
For any linear transformation we know that $T(\mathbf{0})=\mathbf{0}$. Let us assume that the statement is false, that is $T(\mathbf{x})=\mathbf{0}$ has only the trivial solution, but $T$ is not injective. IF $T$ is not injective there exist a point $\mathbf{y}$ in the codomain that is the image of two points in the domain

$$
\begin{aligned}
& T\left(\mathbf{x}_{1}\right)=\mathbf{y} \\
& T\left(\mathbf{x}_{2}\right)=\mathbf{y}
\end{aligned}
$$

If we know subtract the two equations we have

$$
\begin{array}{cl}
T\left(\mathbf{x}_{1}\right)-T\left(\mathbf{x}_{2}\right)=\mathbf{0} & \\
T\left(\mathbf{x}_{1}-\mathbf{x}_{2}\right)=\mathbf{0} & T \text { is linear } \\
\mathbf{x}_{1}-\mathbf{x}_{2}=\mathbf{0} & \text { There is only one solution of } T(\mathbf{x})=\mathbf{0} \\
\mathbf{x}_{1}=\mathbf{x}_{2} & \text { contradiction (q.e.d.) }
\end{array}
$$

## Classification of functions

## Theorem 9.2

Let $T(\mathbf{x})=A \mathbf{x}$ be a linear transformation. Then:
(1) Range $(T)=\mathbb{R}^{m}$ iff $\operatorname{Span}\left(\mathbf{a}_{1}, \mathbf{a}_{2}, \ldots, \mathbf{a}_{n}\right)=\mathbb{R}^{m}$.
(2) $T$ is injective iff all columns of $A$ are linearly independent.

Proof
(1) According to Theorem 3.2, the columns of $A$ span $\mathbb{R}^{m}$ if for each $\mathbf{b} \in \mathbb{R}^{m}$, the equation $A \mathbf{x}=\mathbf{b}$ is consistent, that is, if there exists at least one solution of $T(\mathbf{x})=\mathbf{b}$. If this is true, then Range $(T)=\mathbb{R}^{m}$.
(2) According to Theorem 9.1, $T$ is injective iff $T(\mathbf{x})=\mathbf{0}$ has only the trivial solution, or what is the same iff $A \mathbf{x}=\mathbf{0}$ has only the trivial solution. This happens only if the columns of $A$ are linearly independent as stated by Theorem 6.1.

## Classification of functions

## Example

Let $T(\mathbf{x})=\left(\begin{array}{c}3 x_{1}+x_{2} \\ 5 x_{1}+7 x_{2} \\ x_{1}+3 x_{2}\end{array}\right)$ :
(1) Show that it is a linear transformation
(2) Does it map $\mathbb{R}^{2}$ onto $\mathbb{R}^{3}$ ?

## Solution

(1) The transformation is of the form $T(\mathbf{x})=A \mathbf{x}$ with $A=\left(\begin{array}{ll}3 & 1 \\ 5 & 7 \\ 1 & 3\end{array}\right)$ and, therefore, the transformation is linear.
(2) The columns of $A$ are linearly independent (because they are not multiples of each other), then, by the previous theorem, the transformation is injective. However, they do not span all $\mathbb{R}^{3}$ (since they are only two vectors and for spanning all $\mathbb{R}^{3}$ we need at least 3 vectors). Consequently, the transformation is not surjective, and it does not map $\mathbb{R}^{2}$ onto $\mathbb{R}^{3}$.

## Exercises

## Exercises

From Lay (4th ed.), Chapter 1, Section 9:

- 1.9.1
- 1.9.3
- 1.9.17
- 1.9.33
- 1.9.36
- 1.9.37
- 1.9.39


## Outline

(2) Linear equation system

- Introduction (a)
- Gauss-Jordan algorithm (b)
- Interpretation as a subspace (b)
- Existence and uniqueness of solutions (c)
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## More applications

## Construction of a diet

Given the following nutritional information:

| Amounts (g) Supplied per 100 g of Ingredient |  |  |  | Amounts (g) Supplied by Cambridge Diet in One Day |
| :---: | :---: | :---: | :---: | :---: |
| Nutrient | Nonfat milk | Soy flour | Whey |  |
| Protein | 36 | 51 | 13 | 33 |
| Carbohydrate | 52 | 34 | 74 | 45 |
| Fat | 0 | 7 | 1.1 | 3 |

What is the amount of nonfat milk, soy flour and whey needed to provide the protein carbohydrate and fat planned for one day?
Solution
$\left(\begin{array}{rrr|r}36 & 51 & 13 & 33 \\ 52 & 34 & 74 & 45 \\ 0 & 7 & 11 & 3\end{array}\right) \sim\left(\begin{array}{lll|l}1 & 0 & 0 & 0.277 \\ 0 & 1 & 0 & 0.392 \\ 0 & 0 & 1 & 0.233\end{array}\right)$

That is, we need $x_{1}=0.277 \cdot 100 \mathrm{~g}=277 \mathrm{~g}$ of non-fat milk, $x_{2}=392 \mathrm{~g}$ of soy flour and $x_{3}=233 \mathrm{~g}$ of whey.

## More applications

## Dynamic systems: difference equations

In a simplistic model red blood cells (erythrocytes) are created in the bone marrow, then some of them pass to the blood. After some time, old red blood cells are destroyed in the spleen (bazo).


## More applications

## Dynamic systems: difference equations (continued)

Let's say that at every time interval:

- $5 \%$ of the erythrocytes in the marrow leave to the blood stream.
- $2 \%$ of the erythrocytes in the blood stream are destroyed by the spleen.
- 1 M new red blood cells are created at the marrow.

The following equation can be used to determine the amount of erythrocytes at any moment

$$
\binom{x_{\text {marrow }}^{(k+1)}}{x_{\text {blood }}^{(k+1)}}=\left(\begin{array}{cc}
0.95 & 0 \\
0.05 & 0.98
\end{array}\right)\binom{x_{\text {marrow }}^{(k)}}{x_{\text {blood }}^{(k)}}+\binom{10^{6}}{0}
$$

This kind of models is called difference equations.

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# Chapter 3. Matrix algebra 

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## Outline

(3) Matrix algebra

- Matrix operations (a)
- Inverse of a matrix (b)
- Elementary matrices (b)
- An algorithm to invert matrices (b)
- Characterization of invertible matrices (c)
- Invertible linear transformations (c)
- Partitioned matrices (c)
- LU factorization (d)
- An application to computer graphics and image processing (d)
- Subspaces of $\mathbb{R}^{n}(\mathrm{e})$
- Dimension and rank (e)


## References


D. Lay. Linear algebra and its applications (3rd ed). Pearson (2006). Chapter 2.

## A little bit of history

Matrices appeared as a regular arrangement of numbers more than 2,000 years ago. However, it was during the $\mathrm{XVII}{ }^{\text {th }}, \mathrm{XVIII}{ }^{\text {th }}$ and $\mathrm{XIX}^{\text {th }}$ centuries that they developed in the way we know them now. Some important names in their modern development are Seki Takakazu (1683), Gottfried Leibniz (1693), Gabriel Cramer (1750), James Sylvester (1850), and Arthur Cayley (1858). They were applied in all kind of mathematical problems as a way to organize calculations.


To know more about the history of matrix algebra visit

- http://www-groups.dcs.st-and.ac.uk/~history/PrintHT/Matrices_ and_determinants.html


## Applications

Finite elements has been one of the most successful approaches to biomechanical modeling. In the figure we show one of such a model for the heart. Using this model, all kind of local stresses can be calculated.

J. Berkley, S. Weghorst, H. Gladstone, G. Raugi, D. Berg, M. Ganter. Banded Matrix Approach to Finite Element Modeling for Soft Tissue Simulation.

## Outline

(3) Matrix algebra

- Matrix operations (a)
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## Basic definitions

## Definition 1.1 (Matrix)

Informally, we can define a matrix as a regular arrangement of numbers that are laid out in a grid of $m$ rows and $n$ columns. More formally, we say that $A \in \mathcal{M}_{m \times n}$. We denote as $\mathbf{a}_{j}$ as its $j$-th column, and $a_{i j}$ the element in the $i$-th row and the $j$-th column.

$$
A=\left(\begin{array}{llll}
\mathbf{a}_{1} & \mathbf{a}_{2} & \ldots & \mathbf{a}_{n}
\end{array}\right)=\left(\begin{array}{cccc}
a_{11} & a_{12} & \ldots & a_{1 n} \\
a_{21} & a_{22} & \ldots & a_{2 n} \\
\ldots & \ldots & \ldots & \ldots \\
a_{m 1} & a_{m 2} & \ldots & a_{m n}
\end{array}\right)
$$

The main diagonal is the vector given by $\left(a_{11}, a_{22}, \ldots\right)$. Two important special matrices are the identity matrix $\left(I \in \mathcal{M}_{n \times n}\right)$ that is zero everywhere except the main diagonal that is full of 1 s ; and the zero matrix ( $0 \in \mathcal{M}_{m \times n}$ ) that is zero everywhere.

## Example <br> MATLAB: A=[112 3; 4 5 6]

## Matrix operations

## Definition 1.2 (Sum with a scalar)

We define the sum operator between a scalar and a matrix as:

$$
\left.\begin{aligned}
+: \mathbb{R} \times \mathcal{M}_{m \times n} & \rightarrow \quad \mathcal{M}_{m \times n} \\
& +(k, A)
\end{aligned} \quad \rightarrow B=k+A \quad \right\rvert\, b_{i j}=k+a_{i j} j
$$

We overload the notation to define the sum operator between a matrix and a scalar as

$$
\left.\begin{aligned}
+: \mathcal{M}_{m \times n} \times \mathbb{R} & \rightarrow \quad \mathcal{M}_{m \times n} \\
& +(A, k)
\end{aligned} \quad \rightarrow B=A+k \quad \right\rvert\, b_{i j}=a_{i j}+k
$$

## Example

$$
\begin{aligned}
& A=\left(\begin{array}{ccc}
1 & 2 & 3 \\
-1 & -2 & -3
\end{array}\right) \\
& B=1+A=\left(\begin{array}{ccc}
2 & 3 & 4 \\
0 & -1 & -2
\end{array}\right) \\
& \text { MATLAB: B }=1+\mathrm{A}
\end{aligned}
$$

## Properties

$$
\begin{aligned}
k+A & =A+k \\
\left(k_{1}+k_{2}\right)+A & =k_{1}+\left(k_{2}+A\right)
\end{aligned}
$$

## Matrix operations

## Definition 1.3 (Multiplication with a scalar)

We define the multiplication operator between a scalar and a matrix as:

$$
\left.\begin{aligned}
\cdot: \mathbb{R} \times \mathcal{M}_{m \times n} & \rightarrow \quad \mathcal{M}_{m \times n} \\
& \cdot(k, A)
\end{aligned} \quad \rightarrow B=k+A \quad \right\rvert\, b_{i j}=k a_{i j}
$$

We overload the notation to define the multiplication operator between a matrix and a scalar as

$$
\begin{aligned}
\therefore \mathcal{M}_{m \times n} \times \mathbb{R} & \rightarrow \mathcal{M}_{m \times n} \\
\cdot(A, k) & \rightarrow B=A k \quad \mid b_{i j}=a_{i j} k
\end{aligned}
$$

## Example

$$
\begin{aligned}
A & =\left(\begin{array}{ccc}
1 & 2 & 3 \\
-1 & -2 & -3 \\
2 & 4 & 6 \\
-2 & -4 & -6
\end{array}\right) \\
B=2 A & =\left(\begin{array}{cl}
\end{array}\right) \\
\text { MATLAB: } & =2 * \mathrm{~A}
\end{aligned}
$$

Properties

$$
\begin{aligned}
k A & =A k \\
\left(k_{1} k_{2}\right) A & =k_{1}\left(k_{2} A\right) \\
\left(k_{1}+k_{2}\right) A & =k_{1} A+k_{2} A
\end{aligned}
$$

## Matrix operations

## Definition 1.4 (Sum of two matrices)

We define the sum operator between two matrices as:

$$
\begin{array}{rll}
+: \mathcal{M}_{m \times n} \times \mathcal{M}_{m \times n} & \rightarrow \mathcal{M}_{m \times n} \\
+(A, B) & \rightarrow C=A+B \quad \mid c_{i j}=a_{i j}+b_{i j}
\end{array}
$$

Example

$$
A=\left(\begin{array}{ccc}
1 & 2 & 3 \\
-1 & -2 & -3
\end{array}\right)
$$

Properties

$$
\begin{aligned}
A+B & =B+A \\
A+(B+C) & =(A+B)+C \\
A+0 & =A \\
k(A+B) & =k A+k B
\end{aligned}
$$

MATLAB: $\mathrm{C}=\mathrm{A}+\mathrm{B}$

## Matrix operations

## Proof of the properties

We are not proving all properties, although all of them follow the same strategy. Let's see an example

$$
k(A+B)=k A+k B
$$

Proof
Let us develop the left hand side

$$
\begin{array}{r|l}
C=A+B & \begin{array}{l}
c_{i j}=a_{i j}+b_{i j} \\
D=k C=k(A+B)
\end{array} \\
d_{i j}=k c_{i j}=k\left(a_{i j}+b_{i j}\right)=k a_{i j}+k b_{i j}
\end{array}
$$

Now, the right hand side

$$
\begin{array}{r|l}
E=k A & e_{i j}=k a_{i j} \\
F=k B & f_{i j}=k b_{i j} \\
G=E+F=k A+k B & g_{i j}=e_{i j}+f_{i j}=k a_{i j}+k b_{i j}
\end{array}
$$

It is obvious that $d_{i j}=g_{i j}$, and consequently $k(A+B)=k A+k B$. (q.e.d.)

## Matrix operations

## Definition 1.5 (Multiplication of two matrices)

We define the multiplication operator between two matrices as:

$$
\begin{aligned}
\cdot \mathcal{M}_{m \times n} \times \mathcal{M}_{n \times p} & \rightarrow \mathcal{M}_{m \times p} \\
\cdot(A, B) & \rightarrow C=A B \quad \mid c_{i j}=\sum_{k=1}^{n} a_{i k} b_{k j}
\end{aligned}
$$

If we consider the different columns of $B$, then we have

$$
B=\left(\begin{array}{llll}
\mathbf{b}_{1} & \mathbf{b}_{2} & \ldots & \mathbf{b}_{p}
\end{array}\right) \Rightarrow A B=\left(\begin{array}{llll}
A \mathbf{b}_{1} & A \mathbf{b}_{2} & \ldots & A \mathbf{b}_{p}
\end{array}\right)
$$

That can be interpreted as "the $j$-th column of $A B$ is a weighted sum of the columns of matrix $A$ using the weights defined by the $j$-th column of $B$ ".

## Example

MATLAB: A*B

## Matrix operations

## Example

Let $A=\left(\begin{array}{cc}2 & 3 \\ 1 & -5\end{array}\right)$ and $B=\left(\begin{array}{ccc}4 & 3 & 6 \\ 1 & -2 & 3\end{array}\right)$. Then,
$A \mathbf{b}_{1}=\left(\begin{array}{cc}2 & 3 \\ 1 & -5\end{array}\right)\binom{4}{1}=\binom{11}{-1}$
$A \mathbf{b}_{2}=\left(\begin{array}{cc}2 & 3 \\ 1 & -5\end{array}\right)\binom{3}{-2}=\binom{0}{13}$
$A \mathbf{b}_{3}=\left(\begin{array}{cc}2 & 3 \\ 1 & -5\end{array}\right)\binom{6}{3}=\binom{21}{-9}$
$A B=\left(\begin{array}{lll}A \mathbf{b}_{1} & A \mathbf{b}_{2} & A \mathbf{b}_{3}\end{array}\right)=\left(\begin{array}{ccc}11 & 0 & 21 \\ -1 & 13 & -9\end{array}\right)$
To directly compute a specific entry, for instance, $(A B)_{23}$ we have to multiply the 2nd row of $A$ and the third column of $B$

$$
(A B)_{23}=\left[\left(\begin{array}{cc}
2 & 3 \\
1 & -5
\end{array}\right)\left(\begin{array}{ccc}
4 & 3 & 6 \\
1 & -2 & 3
\end{array}\right)\right]=1 \cdot 6+(-5) \cdot 3=-9
$$

## Matrix operations

## Geometrical interpretation

Consider the linear transformations

$$
\begin{aligned}
& T_{A}(\mathbf{x})=A \mathbf{x} \\
& T_{B}(\mathbf{x})=B \mathbf{x}
\end{aligned}
$$

that map any input vector using the matrix $A$ or $B$, respectively. Now consider the sequential application of first $T_{B}$, and then $T_{A}$, as shown in the following figure:


Matrix multiplication helps us to define a single transformation such that we can transform the original vectors in a single step:

$$
T_{A B}(\mathbf{x})=(A B) \mathbf{x}=A(B \mathbf{x})=T_{A}\left(T_{B}(\mathbf{x})\right)
$$

## Matrix operations

## Property

$\operatorname{row}_{i}(A B)=\operatorname{row}_{i}(A) B$

## Example (continued)

$$
\operatorname{row}_{1}(A B)=\operatorname{row}_{1}(A) B=\left(\begin{array}{ll}
2 & 3
\end{array}\right)\left(\begin{array}{ccc}
4 & 3 & 6 \\
1 & -2 & 3
\end{array}\right)=\left(\begin{array}{lll}
11 & 0 & 21
\end{array}\right)
$$

More properties

$$
\begin{array}{cl}
A(B C)=(A B) C & \text { Associativity } \\
A(B+C)=A B+A C & \text { Left distributivity } \\
(A+B) C=A C+B C & \text { Right distributivity } \\
r(A B)=(r A) B=A(r B) & \text { For any scalar } r \\
I_{m} A=A=A I_{n} & \text { For } A \in \mathcal{M}_{m \times n}
\end{array}
$$

## Matrix operations

Proof $A(B C)=(A B) C$
Let us consider the column decomposition of matrix $C$.

$$
\begin{gathered}
C=\left(\begin{array}{llll}
\mathbf{c}_{1} & \mathbf{c}_{2} & \ldots & \mathbf{c}_{p}
\end{array}\right) \Rightarrow \\
B C=\left(\begin{array}{llll}
B \mathbf{c}_{1} & B \mathbf{c}_{2} & \ldots & B \mathbf{c}_{p}
\end{array}\right) \Rightarrow \\
A(B C)=\left(\begin{array}{llll}
A\left(B \mathbf{c}_{1}\right) & A\left(B \mathbf{c}_{2}\right) & \ldots & A\left(B \mathbf{c}_{p}\right)
\end{array}\right)
\end{gathered}
$$

But we have seen in the geometrical interpretation of matrix multiplication that $A\left(B \mathbf{c}_{i}\right)=(A B) \mathbf{c}_{i}$, therefore

$$
A(B C)=\left((A B) \mathbf{c}_{1} \quad(A B) \mathbf{c}_{2} \quad \ldots \quad(A B) \mathbf{c}_{p}\right)=(A B) C
$$

## Warnings

- $A B \neq B A$, matrix multiplication is not commutative.
- $A B=A C \nRightarrow B=C$.
- $A B=0 \nRightarrow B=0$ or $C=0$.


## Matrix operations

## Definition 1.6 (Power of a matrix)

If $A \in \mathcal{M}_{n \times n}$, then the $k$-th power of the matrix is defined as

$$
A^{k}=\underbrace{A \cdot A \cdot \ldots \cdot A}_{k \text { times }}
$$

Note: $A^{0}=I_{n}$

## Example

MATLAB: $\mathrm{A}^{\wedge} \mathrm{k}$

## Matrix operations

## Definition 1.7 (Transpose)

If $A \in \mathcal{M}_{m \times n}$, then the transpose of $A\left(A^{T}\right)$ is a matrix in $\mathcal{M}_{n \times m}$ such that the rows of $A$ are the columns of $A^{T}$, or more formally

$$
\left(A^{T}\right)_{i j}=A_{j i}
$$

## Example

$$
\begin{gathered}
A=\left(\begin{array}{lll}
1 & 2 & 3 \\
4 & 5 & 6
\end{array}\right) \Rightarrow A^{T}=\left(\begin{array}{ll}
1 & 4 \\
2 & 5 \\
3 & 6
\end{array}\right) \\
\text { MATLAB: A }
\end{gathered}
$$

## Properties

$$
\begin{gathered}
\left(A^{T}\right)^{T}=A \\
(A+B)^{T}=A^{T}+B^{T} \\
(r A)^{T}=r A^{T} \\
(A B)^{T}=B^{T} A^{T}
\end{gathered}
$$

## Matrix operations

## $\operatorname{Proof}(A B)^{T}=B^{T} A^{T}$

Let $A \in \mathcal{M}_{m \times n}$ and $B \in \mathcal{M}_{n \times p}$ By the definition of matrix multiplication we know that

$$
(A B)_{i j}=\sum_{k=1}^{n} a_{i k} b_{k j}
$$

Let $B^{\prime}=B^{T}$ and $A^{\prime}=A^{T}$. For the same reason

$$
\left(B^{T} A^{T}\right)_{i j}=\left(B^{\prime} A^{\prime}\right)_{i j}=\sum_{k=1}^{n} b_{i k}^{\prime} a_{k j}^{\prime}
$$

But $b_{i k}^{\prime}=b_{k i}$ and $a_{k j}^{\prime}=a_{j k}$, consequently

$$
\left(B^{T} A^{T}\right)_{i j}=\sum_{k=1}^{n} b_{k i} a_{j k}=\sum_{k=1}^{n} a_{j k} b_{k i}=(A B)_{j i}
$$

or what is the same

$$
B^{T} A^{T}=(A B)^{T}
$$

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 2, Section 1:

- 2.1.3
- 2.1.23
- 2.1.10
- 2.1.24
- 2.1.12
- 2.1.18
- 2.1.19
- 2.1.20
- 2.1.22
- 2.1.25
- 2.1.26
- 2.1.27
- 2.1.39 (bring computer)
- 2.1.40 (bring computer)


## Outline

(3) Matrix algebra

- Matrix operations (a)
- Inverse of a matrix (b)
- Elementary matrices (b)
- An algorithm to invert matrices (b)
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- Dimension and rank (e)


## Matrix inverse

## Example

The inverse of a number is a clear concept

$$
5 \frac{1}{5}=5 \cdot 5^{-1}=1=5^{-1} \cdot 5
$$

## Definition 2.1 (Inverse of a matrix)

A matrix $A \in \mathcal{M}_{n \times n}$ is invertible (or non-singular) if there exists another matrix $C \in \mathcal{M}_{n \times n}$ such that $A C=I_{n}=C A$. $C$ is called the inverse of $A$ and it is denoted as $A^{-1}$. If $A$ is not invertible, it is said to be singular. (MATLAB: $\operatorname{inv}(A)$ )

## Properties

The inverse of a matrix is unique.
Proof
Let us assume that there exist two different inverses $C_{1}$ and $C_{2}$. Then,

$$
C_{2}=C_{2} I=C_{2}\left(A C_{1}\right)=\left(C_{2} A\right) C_{1}=I C_{1}=C_{1}
$$

which is a contradiction and, therefore, the inverse must be unique. (q.e.d.)

## Matrix inverse

## Example

$$
\text { Let } A=\left(\begin{array}{cc}
2 & 5 \\
-3 & -7
\end{array}\right) \text { and } A^{-1}=\left(\begin{array}{cc}
-7 & -5 \\
3 & 2
\end{array}\right)
$$

It can easily be verified that

$$
A A^{-1}=A^{-1} A=I_{2}=\left(\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right)
$$

Theorem 2.1 (Inverse of a $2 \times 2$ matrix)
Let $A=\left(\begin{array}{ll}a & b \\ c & d\end{array}\right)$. If $a d-b c \neq 0$, then $A$ is invertible and its inverse is

$$
A^{-1}=\frac{1}{a d-b c}\left(\begin{array}{cc}
d & -b \\
-c & a
\end{array}\right)
$$

Proof It is easy to verify that $A A^{-1}=A^{-1} A=I_{2}$.

## Matrix inverse

## Theorem 2.2

If $A \in \mathcal{M}_{n \times n}$ is invertible, then for every $b \in \mathbb{R}^{n}$, the equation $A \mathbf{x}=\mathbf{b}$ has a unique solution that is $\mathbf{x}=A^{-1} \mathbf{b}$.
Proof
Proof $\mathbf{x}=A^{-1} \mathbf{b}$ is a solution
If we substitute the solution in the equation we have

$$
A \mathbf{x}=A\left(A^{-1} \mathbf{b}\right)=\left(A A^{-1}\right) \mathbf{b}=\mathbf{b} \text { (q.e.d.) }
$$

Proof $\mathbf{x}=A^{-1} \mathbf{b}$ is the unique solution
Let us assume that $\mathbf{x}^{\prime} \neq \mathbf{x}$ is also a solution, then

$$
A \mathbf{x}^{\prime}=\mathbf{b}
$$

If we multiply on the left by $A^{-1}$, we have

$$
A^{-1} A \mathbf{x}^{\prime}=A^{-1} \mathbf{b} \Rightarrow \mathbf{x}^{\prime}=\mathbf{x}
$$

which is obviously a contradiction and, therefore, $\mathbf{x}=A^{-1} \mathbf{b}$ must be the unique solution. (q.e.d.)

## Matrix inverse

## Example

EXAMPLE 3 A horizontal elastic beam is supported at each end and is subjected to forces at points $1,2,3$, as shown in Fig. 1. Let $\mathbf{f}$ in $\mathbb{R}^{3}$ list the forces at these points, and let $\mathbf{y}$ in $\mathbb{R}^{3}$ list the amounts of deflection (that is, movement) of the beam at the three points. Using Hooke's law from physics, it can be shown that

$$
\mathbf{y}=D \mathbf{f}
$$

where $D$ is a flexibility matrix. Its inverse is called the stiffness matrix. Describe the physical significance of the columns of $D$ and $D^{-1}$.


FIGURE 1 Deflection of an elastic beam.

## Matrix inverse

## Example (continued)

Consider the equation $\mathbf{y}=D \mathbf{f}, D=\left(\begin{array}{ccc}1 & \frac{1}{2} & \frac{1}{4} \\ \frac{1}{2} & 1 & \frac{1}{2} \\ \frac{1}{4} & \frac{1}{2} & 1\end{array}\right)$ and the fact that

$$
D=D I=\left(\begin{array}{lll}
D \mathbf{e}_{1} & D \mathbf{e}_{2} & D \mathbf{e}_{3}
\end{array}\right)
$$

Therefore, the $i$-th column of $D$ can be interpreted as the deflection at the different points when a unit force $\left(\mathbf{e}_{i}\right)$ is applied onto the $i$-th point. In our example when we apply a unit force at point 1 , the first column of $D$ is $\left(1, \frac{1}{2}, \frac{1}{4}\right)$ meaning that the first point displaces 1 m ., the second point $\frac{1}{2} \mathrm{~m}$., and the third point $\frac{1}{4} \mathrm{~m}$.

## Matrix inverse

## Example (continued)

If we now consider that $\mathbf{f}=D^{-1} \mathbf{y}, D^{-1}=\left(\begin{array}{ccc}\frac{4}{3} & -\frac{2}{3} & 0 \\ -\frac{2}{3} & \frac{4}{3} & -\frac{2}{3} \\ 0 & -\frac{2}{3} & \frac{4}{3}\end{array}\right)$ and the fact that

$$
D^{-1}=D^{-1} I=\left(\begin{array}{lll}
D^{-1} \mathbf{e}_{1} & D^{-1} \mathbf{e}_{2} & D^{-1} \mathbf{e}_{3}
\end{array}\right)
$$

Therefore, the $i$-th column of $D^{-1}$ can be interpreted as the forces needed to be applied at the different points to produce a unit deformation $\left(\mathbf{e}_{i}\right)$ at the $i$-th point. In our example, to produce a displacement of 1 m . in the first point and none at the other points $\left(\mathbf{e}_{1}=(1,0,0)\right.$, we need to push point 1 with a force of $\frac{4}{3}$ N., to pull point 2 with a force of $-\frac{2}{3} \mathrm{~N}$., and we do not need to apply any force at point 3.

## Matrix inverse

## Theorem 2.3

(1) If $A$ is invertible, then $A^{-1}$ is also invertible and its inverse is $A$.
(2) If $A$ and $B$ are invertible, then $A B$ is also invertible and its inverse is $B^{-1} A^{-1}$
(3) If $A$ is invertible, then $A^{T}$ is also invertible and its inverse is $\left(A^{-1}\right)^{T}$.

Proof 1)
The definition of $A^{-1}$ is that it is a matrix such that

$$
A A^{-1}=A^{-1} A=I
$$

The inverse of $A^{-1}$ must be a matrix $C$ such that

$$
C A^{-1}=A^{-1} C=1
$$

If we compare this equation with the previous one, we easily see that $C=A$ is the inverse of $A^{-1}$.

## Matrix inverse

## Proof 2)

Let us check that $B^{-1} A^{-1}$ is actually the inverse of $A B$

$$
\begin{aligned}
& (A B)\left(B^{-1} A^{-1}\right)=A\left(B B^{-1}\right) A^{-1}=A I A^{-1}=A A^{-1}=I \\
& \left(B^{-1} A^{-1}\right)(A B)=B^{-1}\left(A^{-1} A\right) B=B^{-1} I B=B^{-1} B=I
\end{aligned}
$$

## Proof 3)

Let us check that $\left(A^{-1}\right)^{T}$ is actually the inverse of $A^{T}$

$$
\begin{aligned}
& A^{T}\left(A^{-1}\right)^{T}=\left[(A B)^{T}=B^{T} A^{T}\right]=\left(A^{-1} A\right)^{T}=I^{T}=I \\
& \left(A^{-1}\right)^{T} A^{T}=\left[(A B)^{T}=B^{T} A^{T}\right]=\left(A A^{-1}\right)^{T}=I^{T}=I
\end{aligned}
$$

## Theorem 2.4

We may generalize the previous theorem and state that

$$
\left(A_{1} A_{2} \ldots A_{p}\right)^{-1}=A_{p}^{-1} A_{p-1}^{-1} \ldots A_{2}^{-1} A_{1}^{-1}
$$

## Matrix inverse

## Proof

Let's prove it by weak induction. That is, we know that the statement is true for $p=2$ (by the previous theorem). Let us assume it is true for $p-1$, that is

$$
\left(A_{1} A_{2} \ldots A_{p-1}\right)^{-1}=A_{p-1}^{-1} \ldots A_{2}^{-1} A_{1}^{-1}
$$

We wonder if it is also true for $p$. Let us define $B=A_{1} A_{2} \ldots A_{p-1}$. Then, we can rewrite the left hand side of the theorem as

$$
\left(A_{1} A_{2} \ldots A_{p}\right)^{-1}=\left(B A_{p}\right)^{-1}
$$

This is the inverse of the multiplication of two matrices. We know by the previous theorem that $\left(B A_{p}\right)^{-1}=A_{p}^{-1} B^{-1}$ But we presumed that

$$
B^{-1}=\left(A_{1} A_{2} \ldots A_{p-1}\right)^{-1}=A_{p-1}^{-1} \ldots A_{2}^{-1} A_{1}^{-1}
$$

And consequently

$$
\left(B A_{p}\right)^{-1}=A_{p}^{-1} A_{p-1}^{-1} \ldots A_{2}^{-1} A_{1}^{-1} \text { (q.e.d.) }
$$

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## Elementary matrices

The elementary operations we can perform on the rows of a matrix are
(1) Multiply by a scalar
(2) Swap two rows
(3) Replace a row by a linear combination of two or several rows

All these operations can be represented as matrix multiplications.

## Example

Consider the matrix $A=\left(\begin{array}{lll}a & b & c \\ d & e & f \\ g & h & i\end{array}\right)$
(1) We can multiply the third row by a scalar $r$ by multiplying on the left by the matrix

$$
E_{1}=\left(\begin{array}{lll}
1 & 0 & 0 \\
0 & 1 & 0 \\
0 & 0 & r
\end{array}\right) \Rightarrow E_{1} A=\left(\begin{array}{ccc}
a & b & c \\
d & e & f \\
r g & r h & r i
\end{array}\right)
$$

## Elementary matrices

## Example (continued)

(2) We can swap the first and second rows of the matrix by multiplying on the left by the matrix

$$
E_{2}=\left(\begin{array}{lll}
0 & 1 & 0 \\
1 & 0 & 0 \\
0 & 0 & 1
\end{array}\right) \Rightarrow E_{2} A=\left(\begin{array}{lll}
d & e & f \\
a & b & c \\
g & h & i
\end{array}\right)
$$

(3) We can substitute the third row by $\mathbf{r}_{3}+k_{1} \mathbf{r}_{1}$ by multiplying on the left by the matrix

$$
E_{3}=\left(\begin{array}{ccc}
1 & 0 & 0 \\
0 & 1 & 0 \\
k_{1} & 0 & 1
\end{array}\right) \Rightarrow E_{3} A=\left(\begin{array}{ccc}
a & b & c \\
d & e & f \\
g+k_{1} a & h+k_{1} b & i+k_{1} c
\end{array}\right)
$$

## Definition 3.1 (Elementary matrix)

An elementary matrix is one that differs from the identity matrix by one single, elementary row operation.

## Elementary matrices

## Theorem 3.1

The inverse of an elementary matrix is another elementary matrix of the same type. That is, row operations can be undone.

## Example (continued)

(1) $E_{1}^{-1}=\left(\begin{array}{ccc}1 & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 0 & \frac{1}{r}\end{array}\right)$
(2) $E_{2}^{-1}=\left(\begin{array}{lll}0 & 1 & 0 \\ 1 & 0 & 0 \\ 0 & 0 & 1\end{array}\right)$
(3) $E_{3}^{-1}=\left(\begin{array}{ccc}1 & 0 & 0 \\ 0 & 1 & 0 \\ -k_{1} & 0 & 1\end{array}\right) \quad \begin{aligned} & \text { MATLAB: } \\ & \text { syms } \mathrm{k} 1 \\ & \mathrm{E} 3=\left[\begin{array}{lllllllll}1 & 0 & 0 ; & 0 & 1 & 0 ; & \mathrm{k} 1 & 0 & 1\end{array}\right] ; \\ & \operatorname{inv}(\mathrm{E} 3)\end{aligned}$

## Elementary matrices

## Theorem 3.2

A matrix $A \in \mathcal{M}_{n \times n}$ is invertible iff it is row-equivalent to $I_{n}$. In this case, the sequence of operations that transforms $A$ into $I_{n}$ is also the one that transforms $I_{n}$ into $A^{-1}$

Proof $\Rightarrow$
If $A$ is invertible, then by theorem 2.2 we know that the equation system $A \mathbf{x}=\mathbf{b}$ has a unique solution for every $\mathbf{b}$. If it has a solution for every $\mathbf{b}$, then it must have a pivot in every row, that must be in the diagonal and, consequently the reduced echelon form of $A$ must be $I_{n}$.
Proof $\Leftarrow$
If $A$ is row-equivalent $I_{n}$, then there exists a sequence of elementary matrices that transform $A$ into $I_{n}$

$$
A \sim E_{1} A \sim E_{2} E_{1} A \sim \ldots \sim E_{n} E_{n-1} \ldots E_{2} E_{1} A=I_{n}
$$

$E=E_{n} E_{n-1} \ldots E_{2} E_{1}$ is a candidate to be the inverse of $A$. Since each of the elementary matrices is invertible, and the product of invertible matrices is invertible, then $E$ is invertible and $A$ must be its (unique) inverse. Conversely, $E$ is the inverse of $A$ and $A$ is invertible.

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## An algorithm to invert matrices

## Algorithm

Algorithm: Reduce the augmented matrix $(A \mid I)$
If $A$ is invertible, then $(A \mid I) \sim\left(I \mid A^{-1}\right)$.
If $A$ is not invertible, then we will not be able to reduce $A$ into $I$.
This algorithm is very much used in practice because it is numerically stable and rather efficient.

## Example

Let $A=\left(\begin{array}{ccc}0 & 1 & 2 \\ 1 & 0 & 3 \\ 4 & -3 & 8\end{array}\right)$.
We construct the augmented matrix

$$
\left(\begin{array}{ccc|ccc}
0 & 1 & 2 & 1 & 0 & 0 \\
1 & 0 & 3 & 0 & 1 & 0 \\
4 & -3 & 8 & 0 & 0 & 1
\end{array}\right)
$$

## An algorithm to invert matrices

## Example (continued)

And now we transform it

$$
\begin{aligned}
& \mathbf{r}_{1} \leftrightarrow \mathbf{r}_{2} \\
& \mathbf{r}_{3} \leftarrow \mathbf{r}_{3}-4 \mathbf{r}_{1}\left(\begin{array}{ccc|ccc}
0 & 1 & 2 & 1 & 0 & 0 \\
1 & 0 & 3 & 0 & 1 & 0 \\
4 & -3 & 8 & 0 & 0 & 1 \\
1 & 0 & 3 & 0 & 1 & 0 \\
0 & 1 & 2 & 1 & 0 & 0 \\
4 & -3 & 8 & 0 & 0 & 1
\end{array}\right) \\
& \mathbf{r}_{3} \leftarrow \mathbf{r}_{3}+3 \mathbf{r}_{2}\left(\begin{array}{ccc|ccc}
1 & 0 & 3 & 0 & 1 & 0 \\
0 & 1 & 2 & 1 & 0 & 0 \\
0 & -3 & -4 & 0 & -4 & 1
\end{array}\right) \\
& \mathbf{r}_{3} \leftarrow \frac{1}{2} \mathbf{r}_{3}\left(\begin{array}{ccc|ccc}
1 & 0 & 3 & 0 & 1 & 0 \\
0 & 1 & 2 & 1 & 0 & 0 \\
0 & 0 & 2 & 3 & -4 & 1 \\
1 & 0 & 3 & 0 & 1 & 0 \\
0 & 1 & 2 & 1 & 0 & 0 \\
0 & 0 & 1 & \frac{3}{2} & -2 & \frac{1}{2}
\end{array}\right)
\end{aligned}
$$

## An algorithm to invert matrices

## Example (continued)

$$
\mathbf{r}_{2} \leftarrow \mathbf{r}_{2}-2 \mathbf{r}_{3}\left(\begin{array}{ccc|ccc}
1 & 0 & 3 & 0 & 1 & 0 \\
0 & 1 & 2 & 1 & 0 & 0 \\
0 & 0 & 1 & \frac{3}{2} & -2 & \frac{1}{2}
\end{array}\right)
$$

Since $A$ is row-equivalent to $I_{3}$, then $A$ is invertible and its inverse is $A^{-1}=\left(\begin{array}{ccc}-\frac{9}{2} & 7 & -\frac{3}{2} \\ -2 & 4 & -1 \\ \frac{3}{2} & -2 & \frac{1}{2}\end{array}\right)$. To finalize the exercise we should check that

$$
A A^{-1}=A^{-1} A=I_{3}
$$

## An algorithm to invert matrices

## A new interpretation of matrix inversion

By constructing the augmented matrix $(A \mid I)$ we are simultaneously solving multiple equation systems

$$
A \mathbf{x}=\mathbf{e}_{1} \quad A \mathbf{x}=\mathbf{e}_{2} \quad A \mathbf{x}=\mathbf{e}_{3}
$$

## Example (continued)

$$
\left(\begin{array}{ccc|c}
0 & 1 & 2 & 1 \\
1 & 0 & 3 & 0 \\
4 & -3 & 8 & 0
\end{array}\right) \quad\left(\begin{array}{ccc|c}
0 & 1 & 2 & 0 \\
1 & 0 & 3 & 1 \\
4 & -3 & 8 & 0
\end{array}\right) \quad\left(\begin{array}{ccc|c}
0 & 1 & 2 & 0 \\
1 & 0 & 3 & 0 \\
4 & -3 & 8 & 1
\end{array}\right)
$$

This note is important because if we want to compute only the $i$-th column of $A^{-1}$ it is enough to solve the equation system

$$
A \mathbf{x}=\mathbf{e}_{i}
$$

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 2, Section 2:

- 2.2.7
- 2.2.11
- 2.2.13
- 2.2.17
- 2.2.19
- 2.2.21
- 2.2.25
- 2.2.36


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## Characterization of invertible matrices

## Theorem 5.1 (The invertible matrix theorem)

Let $A \in \mathcal{M}_{n \times n}$. The following statements are equivalent (either they are all true or they are all false):
i. $A$ is invertible.
ii. $A$ is row-equivalent to $I_{n}$.
iii. A has n pivot positions.
iv. $A \mathbf{x}=\mathbf{0}$ only has the trivial solution $\mathbf{x}=\mathbf{0}$.
v. The columns of $A$ are linearly independent.
vi. The transformation $T(\mathbf{x})=A \mathbf{x}$ is injective.
vii. The equation $A \mathbf{x}=\mathbf{b}$ has at least one solution for every $\mathbf{b} \in \mathbb{R}^{n}$.
viii. The columns of $A$ span $\mathbb{R}^{n}$.
ix. The transformation $T(\mathbf{x})=A \mathbf{x}$ maps $\mathbb{R}^{n}$ onto $\mathbb{R}^{n}$.
x. There exists a matrix $C \in \mathcal{M}_{n \times n}$ such that $C A=I_{n}$.
xi. There exists a matrix $D \in \mathcal{M}_{n \times n}$ such that $A D=I_{n}$.
xii. $A^{T}$ is an invertible matrix

## Characterization of invertible matrices

To prove the theorem we will follow the reasoning scheme below:


Proof $i \Rightarrow x$
If i is true, then x is true simply by doing $C=A^{-1}$.
Proof $x \Rightarrow$ iv See Exercise 2.1.23 in Lay.
Proof iv $\Rightarrow$ iii
See Exercise 2.2.23 in Lay.
Proof iii $\Rightarrow$ ii
If iii is true, then the $n$ pivots have to be in the main diagonal and in this case, the reduced echelon form must be $I_{n}$.

## Characterization of invertible matrices

```
Proof ii =>i
If ii is true, then i is true thanks to Theorem 3.2.
Proof i=>xi
If i is true, then xi is true simply by doing D=A -1.
Proof xi }=>\mathrm{ vii
See Exercise 2.1.24 in Lay.
Proof vii =>i
See Exercise 2.2.24 in Lay.
Proof vii }\Leftrightarrow\mathrm{ viii }\Leftrightarrow\mathrm{ ix
See Theorems 3.2 and 8.2 in Chapter 2.
Proof iv }\Leftrightarrowv\Leftrightarrowv
See Theorems 3.2, 5.1 and 8.1 in Chapter 2.
Proof i=>xii
See Theorem 2.3.
Proof i}\Leftarrowxi
See Theorem 2.3 interchanging the roles of A and AT
```

The power of this theorem is that it connects equation systems to invertibility, linear independence and subspace bases.

## Characterization of invertible matrices

Corollary
(1) If $A$ is invertible, then $A \mathbf{x}=\mathbf{b}$ has a unique solution for every $\mathbf{b} \in \mathbb{R}^{n}$.
(2) If $A, B \in \mathcal{M}_{n \times n}$ and $A B=I_{n}$, then $A$ and $B$ are invertible and $B=A^{-1}$ and $A=B^{-1}$.

Watch out that this corollary only applies to square matrices.

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## Invertible linear transformations

Consider the linear transformation

$$
\begin{aligned}
T: \mathbb{R}^{n} & \rightarrow \mathbb{R}^{n} \\
\mathbf{x} & \rightarrow A \mathbf{x}
\end{aligned}
$$

## Definition 6.1 (Inverse transformation)

$T$ is invertible iff there exists $S: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ such that $\forall \mathbf{x} \in \mathbb{R}^{n}$ :

$$
S(T(\mathbf{x}))=\mathbf{x}=T(S(\mathbf{x}))
$$

## Example

$T(\mathbf{x})=\left(\begin{array}{cc}-1 & 0 \\ 0 & 1\end{array}\right) \mathbf{x}$ is invertible and its inverse is $S(\mathbf{x})=\left(\begin{array}{cc}-1 & 0 \\ 0 & 1\end{array}\right) \mathbf{x}$.
Proof

$$
\begin{aligned}
& S(T(\mathbf{x}))=S\left(\left(\begin{array}{cc}
-1 & 0 \\
0 & 1
\end{array}\right) \mathbf{x}\right)=\left(\begin{array}{cc}
-1 & 0 \\
0 & 1
\end{array}\right)\left(\begin{array}{cc}
-1 & 0 \\
0 & 1
\end{array}\right) \mathbf{x}=\mathbf{x} \\
& T(S(\mathbf{x}))=T\left(\left(\begin{array}{cc}
-1 & 0 \\
0 & 1
\end{array}\right) \mathbf{x}\right)=\left(\begin{array}{cc}
-1 & 0 \\
0 & 1
\end{array}\right)\left(\begin{array}{cc}
-1 & 0 \\
0 & 1
\end{array}\right) \mathbf{x}=\mathbf{x}
\end{aligned}
$$

## Invertible linear transformations

## Example

$T(\mathbf{x})=\left(\begin{array}{ll}1 & 0 \\ 0 & 0\end{array}\right) \mathbf{x}$ is not invertible because $T((1,0))=T((1,1))=(1,0)$, so given the "output" $(1,0)$, we cannot recover the input vector that originated this output.

## Theorem 6.1

If $T$ is invertible, then it is surjective.

## Proof

Consider any vector $\mathbf{b} \in \mathbb{R}^{n}$, we can always apply the transformation $S$ to get a new vector $\mathbf{x}=S(\mathbf{b})$. And then, recover $\mathbf{b}$ making use of the fact that $T$ is the inverse of $S$, that is, $\mathbf{b}=T(\mathbf{x})$. In other words, any vector $\mathbf{b}$ is in the range of $T$ and, therefore, $T$ is surjective.

## Invertible linear transformations

## Theorem 6.2

$T$ is invertible iff $A$ is invertible. If $T$ is invertible, then the only function that satisfies the previous definition is

$$
S(\mathbf{x})=A^{-1} \mathbf{x}
$$

Proof $\Rightarrow$
If $T$ is invertible, then it is surjective (see previous Theorem). Then, $A$ is invertible by Theorem 5.1 (items i and ix).
Proof $\Leftarrow$
If $A$ is invertible, then we may construct the linear transformation $S=A^{-1} \mathbf{x} . S$ is an inverse of $T$ since

$$
\begin{gathered}
S(T(\mathbf{x}))=S(A \mathbf{x})=A^{-1}(A \mathbf{x})=\left(A^{-1} A\right) \mathbf{x}=\mathbf{x} \\
T(S(\mathbf{x}))=T\left(A^{-1} \mathbf{x}\right)=A\left(A^{-1} \mathbf{x}\right)=\left(A A^{-1}\right) \mathbf{x}=\mathbf{x}
\end{gathered}
$$

## Invertible linear transformations

## Proof uniqueness

Let us assume that there are two inverses $S_{1}(\mathbf{x})=B_{1} \mathbf{x}$ and $S_{2}(\mathbf{x})=B_{2} \mathbf{x}$ with $B_{1} \neq B_{2}$. Let $\mathbf{v} \in \mathbb{R}^{n}$ and $\mathbf{v}=T(\mathbf{x})$ for some $\mathbf{x} \in \mathbb{R}^{n}$ (since $T$ is invertible and, therefore, surjective, we are guaranteed that there exists at least one such $\mathbf{x}$ ). Now

$$
\left.\begin{array}{l}
S_{1}(\mathbf{v})=B_{1} A \mathbf{x}=\mathbf{x}=B_{1} \mathbf{v} \\
S_{2}(\mathbf{v})=B_{2} A \mathbf{x}=\mathbf{x}=B_{2} \mathbf{v}
\end{array}\right\} \Rightarrow B_{1} \mathbf{v}=B_{2} \mathbf{v}\left[\forall \mathbf{v} \in \mathbb{R}^{n}\right] \Rightarrow B_{1}=B_{2}
$$

which is a contradiction and, consequently, there exists only one inverse (q.e.d.)

## Definition 6.2 (III-conditioned matrix)

Informally, we say that a matrix $A$ is ill-conditioned if it is "nearly singular". In practice, this implies that the equation system $A \mathbf{x}=\mathbf{b}$ may have large variations in the solution ( $\mathbf{x}$ ) when $\mathbf{b}$ varies slightly.

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 2, Section 3:

- 2.3.13
- 2.3.16
- 2.3.17
- 2.3.33
- 2.3.41


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## Partitioned matrices

Partitioned matrices sometimes help us to gain insight into the structure of the problem by identifying blocks within the matrix.

## Example

$$
A=\left(\begin{array}{ccc|cc|c}
3 & 0 & -1 & 5 & 9 & -2 \\
-5 & 2 & 4 & 0 & -3 & 1 \\
\hline-8 & -6 & 3 & 1 & 7 & -4
\end{array}\right)=\left(\begin{array}{c|c|c}
A_{11} & A_{12} & A_{13} \\
\hline A_{21} & A_{22} & A_{23}
\end{array}\right)
$$

$A \in \mathcal{M}_{3 \times 6}$,
$A_{11} \in \mathcal{M}_{2 \times 3}, A_{12} \in \mathcal{M}_{2 \times 2}, A_{13} \in \mathcal{M}_{2 \times 1}$,
$A_{21} \in \mathcal{M}_{1 \times 3}, A_{22} \in \mathcal{M}_{1 \times 2}, A_{23} \in \mathcal{M}_{1 \times 1}$.
MATLAB:

$$
A=\left[\begin{array}{lllllllllllllllll}
3 & 0 & -1 & 5 & 9 & -2 ; & -5 & 2 & 0 & -3 & 1 ; & -8 & -6 & 3 & 1 & 7 & -4
\end{array}\right] ;
$$

$$
\mathrm{A} 11=\mathrm{A}(1: 2,1: 3)
$$

$$
\mathrm{A} 12=\mathrm{A}(1: 2,4: 5)
$$

$$
\mathrm{A} 13=\mathrm{A}(1: 2,6)
$$

$$
\mathrm{A} 21=\mathrm{A}(3,1: 3)
$$

$$
A 22=A(3,4: 5)
$$

A23 $=A(3,6)$

## Partitioned matrices

## Definition 7.1 (Sum of partitioned matrices)

Let $A$ and $B$ be two matrices partitioned in the same way. Then the blocks of $A+B$ are simply the sum of the corresponding blocks.

$$
A+B=\left(\begin{array}{l|l|l} 
& & \\
& A_{i j} & \\
\hline & &
\end{array}\right)+\left(\begin{array}{l|l|l} 
& & \\
\hline & B_{i j} & \\
\hline & &
\end{array}\right)=\left(\begin{array}{llll} 
& & & \\
& A_{i j}+B_{i j} & \\
\hline & &
\end{array}\right)
$$

## Definition 7.2 (Multiplication by scalar)

The multiplication by a scalar simply multiplies each one of the blocks independently

$$
r A=r\left(\begin{array}{l|l|l} 
& & \\
& A_{i j} & \\
\hline & &
\end{array}\right)=\left(\begin{array}{l|l|l} 
& & \\
\hline & r A_{i j} & \\
\hline & &
\end{array}\right)
$$

## Partitioned matrices

## Definition 7.3 (Multiplication of partitioned matrices)

Multiply the different block as if they were scalars (but applying matrix multiplication).

## Example

Let $A=\left(\begin{array}{ccc|cc}2 & -3 & 1 & 0 & -4 \\ 1 & 5 & -2 & 3 & -1 \\ \hline 0 & -4 & -2 & 7 & -1\end{array}\right)=\left(\begin{array}{c|c}A_{11} & A_{12} \\ \hline A_{21} & A_{22}\end{array}\right)$
and $B=\left(\begin{array}{cc}6 & 4 \\ -2 & 1 \\ -3 & 7 \\ \hline-1 & 3 \\ 5 & 2\end{array}\right)=\left(\frac{B_{1}}{B_{2}}\right)$.
Then, $A B=\left(\begin{array}{ll}A_{11} & A_{12} \\ A_{21} & A_{22}\end{array}\right)\binom{B_{1}}{B_{2}}=\binom{A_{11} B_{1}+A_{12} B_{2}}{A_{21} B_{1}+A_{22} B_{2}}=\left(\begin{array}{cc}-5 & 4 \\ -6 & 2 \\ 2 & 1\end{array}\right)$

## Partitioned matrices

Theorem 7.1 (Multiplication of matrices)
Let $A \in \mathcal{M}_{m \times n}$ and $B \in \mathcal{M}_{n \times p}$, then

$$
A B=\sum_{k=1}^{n} \operatorname{column}_{k}(A) \operatorname{row}_{k}(B)
$$

Proof
Let us analyze each one of the terms in the sum

$$
\begin{gathered}
\operatorname{column}_{k}(A) \operatorname{row}_{k}(B)=\left(\begin{array}{c}
a_{1 k} \\
a_{2 k} \\
\ldots \\
a_{m k}
\end{array}\right)\left(\begin{array}{llll}
b_{k 1} & b_{k 2} & \ldots & b_{k p}
\end{array}\right)= \\
\left(\begin{array}{cccc}
a_{1 k} b_{k 1} & a_{1 k} b_{k 2} & \ldots & a_{1 k} b_{k p} \\
a_{2 k} b_{k 1} & a_{2 k} b_{k 2} & \ldots & a_{2 k} b_{k p} \\
\ldots & \ldots & \ldots & \ldots \\
a_{m k} b_{k 1} & a_{m k} b_{k 2} & \ldots & a_{m k} b_{k p}
\end{array}\right)
\end{gathered}
$$

## Partitioned matrices

In general, the $i j$-th term is

$$
\left(\operatorname{column}_{k}(A) \operatorname{row}_{k}(B)\right)_{i j}=a_{i k} b_{k j}
$$

If we now analyze the $i j$-th element of the sum

$$
\left(\sum_{k=1}^{n} \operatorname{column}_{k}(A) \operatorname{row}_{k}(B)\right)_{i j}=\sum_{k=1}^{n}\left(\operatorname{column}_{k}(A) \operatorname{row}_{k}(B)\right)_{i j}=\sum_{k=1}^{n} a_{i k} b_{k j}
$$

But this is the definition of matrix multiplication and, therefore,

$$
\left(\sum_{k=1}^{n} \operatorname{column}_{k}(A) \operatorname{row}_{k}(B)\right)_{i j}=(A B)_{i j} \text { (q.e.d.) }
$$

## Partitioned matrices

## Definition 7.4 (Transpose of partitioned matrices)

Transpose the partitioned matrix as if it were composed of scalars, and transpose each one of the blocks.

## Example

$$
A=\left(\begin{array}{c|c|c}
A_{11} & A_{12} & A_{13} \\
\hline A_{21} & A_{22} & A_{23} \\
\hline A_{31} & A_{32} & A_{33}
\end{array}\right) \Rightarrow A^{T}=\left(\begin{array}{c|c|c}
A_{11}^{T} & A_{21}^{T} & A_{31}^{T} \\
\hline A_{12}^{T} & A_{22}^{T} & A_{32}^{T} \\
\hline A_{13}^{T} & A_{23}^{T} & A_{33}^{T}
\end{array}\right)
$$

## Example

$$
A=\left(\begin{array}{ccc|cc}
2 & -3 & 1 & 0 & -4 \\
1 & 5 & -2 & 3 & -1 \\
\hline 0 & -4 & -2 & 7 & -1
\end{array}\right) \Rightarrow A^{T}=\left(\begin{array}{cc|c}
2 & 1 & 0 \\
-3 & 5 & -4 \\
1 & -2 & -2 \\
\hline 0 & 3 & 7 \\
-4 & -1 & -1
\end{array}\right)
$$

## Partitioned matrices

## Definition 7.5 (Inverse of partitioned matrices)

The formula for each one of the cases is worked out particularly for that case. Here go a couple of examples.

## Example

Let $A=\left(\begin{array}{c|c|c}A_{11} & 0 & 0 \\ \hline 0 & A_{22} & 0 \\ \hline 0 & 0 & A_{33}\end{array}\right)$.
$A \in \mathcal{M}_{n \times n}, A_{11} \in \mathcal{M}_{p \times p}, A_{22} \in \mathcal{M}_{q \times q}, A_{33} \in \mathcal{M}_{r \times r}$ such that $p+q+r=n$. We look for a matrix $B$ such that

$$
\begin{aligned}
& \left(\begin{array}{c|c|c}
A_{11} & 0 & 0 \\
\hline 0 & A_{22} & 0 \\
\hline 0 & 0 & A_{33}
\end{array}\right)\left(\begin{array}{c|c|c}
B_{11} & B_{12} & B_{13} \\
\hline B_{21} & B_{22} & B_{23} \\
\hline B_{31} & B_{32} & B_{33}
\end{array}\right)=\left(\begin{array}{c|c|c}
I_{p} & 0 & 0 \\
\hline 0 & I_{q} & 0 \\
\hline 0 & 0 & I_{r}
\end{array}\right) \Rightarrow \\
& \left(\begin{array}{c|c|c}
A_{11} B_{11} & A_{11} B_{12} & A_{11} B_{13} \\
\hline A_{22} B_{21} & A_{22} B_{22} & A_{22} B_{23} \\
\hline A_{33} B_{31} & A_{33} B_{32} & A_{33} B_{33}
\end{array}\right)=\left(\begin{array}{c|c|c}
I_{p} & 0 & 0 \\
\hline 0 & I_{q} & 0 \\
\hline 0 & 0 & I_{r}
\end{array}\right)
\end{aligned}
$$

## Partitioned matrices

## Example (continued)

For each one of the entries we have a set of equations:

$$
\begin{aligned}
& \forall A_{11} \in \mathcal{M}_{p \times p} A_{11} B_{11}=I_{p} \Rightarrow B_{11}=A_{11}^{-1} \\
& \forall A_{11} \in \mathcal{M}_{p \times p} A_{11} B_{12}=0 \Rightarrow B_{12}=0 \\
& \forall A_{11} \in \mathcal{M}_{p \times p} A_{11} B_{13}=0 \Rightarrow B_{13}=0 \\
& \forall A_{22} \in \mathcal{M}_{q \times q} A_{22} B_{21}=0 \Rightarrow B_{21}=0 \\
& \forall A_{22} \in \mathcal{M}_{q \times 9} A_{22} B_{22}=I_{q} \Rightarrow B_{22}=A_{22}^{-1} \\
& \forall A_{22} \in \mathcal{M}_{q \times q} A_{22} B_{23}=0 \Rightarrow B_{23}=0 \\
& \forall A_{33} \in \mathcal{M}_{r \times r} A_{33} B_{31}=0 \Rightarrow B_{31}=0 \\
& \forall A_{33} \in \mathcal{M}_{r \times r} A_{33} B_{32}=0 \Rightarrow B_{32}=0 \\
& \forall A_{33} \in \mathcal{M}_{r \times r} A_{33} B_{33}=I_{r} \Rightarrow B_{33}=A_{33}^{-1}
\end{aligned}
$$

Finally,

$$
B=\left(\begin{array}{c|c|c}
A_{11}^{-1} & 0 & 0 \\
\hline 0 & A_{22}^{-1} & 0 \\
\hline 0 & 0 & A_{33}^{-1}
\end{array}\right)
$$

## Partitioned matrices

## Example

Let $A=\left(\begin{array}{c|c}A_{11} & A_{12} \\ \hline 0 & A_{22}\end{array}\right)$.
$A \in \mathcal{M}_{n \times n}, A_{11} \in \mathcal{M}_{p \times p}, A_{12} \in \mathcal{M}_{p \times q}, A_{22} \in \mathcal{M}_{q \times q}$ such that $p+q=n$.
We look for a matrix $B$ such that

$$
\begin{gathered}
=\left(\begin{array}{c|c}
A_{11} & A_{12} \\
\hline 0 & A_{22}
\end{array}\right)\left(\begin{array}{c|c}
B_{11} & B_{12} \\
\hline B_{21} & B_{22}
\end{array}\right)=\left(\begin{array}{c|c|c}
I_{p} & 0 \\
\hline 0 & I_{q}
\end{array}\right) \Rightarrow \\
\left(\begin{array}{c|c|c}
A_{11} B_{11}+A_{12} B_{21} & A_{11} B_{12}+A_{12} B_{22} \\
\hline A_{22} B_{21} & A_{22} B_{22}
\end{array}\right)=\left(\begin{array}{c|c}
I_{p} & 0 \\
\hline 0 & I_{q}
\end{array}\right)
\end{gathered}
$$

## Partitioned matrices

## Example (continued)

For each one of the entries we have a set of equations:

$$
\begin{array}{ll}
\forall A_{22} \in \mathcal{M}_{q \times q} A_{22} B_{21}=0 \Rightarrow B_{21}=0 \\
\forall A_{22} \in \mathcal{M}_{q \times q} A_{22} B_{22}=I_{q} \Rightarrow & B_{22}=A_{22}^{-1} \\
\forall A_{11} \in \mathcal{M}_{q \times q}, A_{12} \in \mathcal{M}_{p \times q} & A_{11} B_{11}+A_{12} B_{21}=I_{p} \Rightarrow\left[B_{21}=0\right] \Rightarrow \\
& A_{11} B_{11}=I_{p} \Rightarrow B_{11}=A_{11}^{-1} \\
\forall A_{11} \in \mathcal{M}_{q \times q}, A_{12} \in \mathcal{M}_{p \times q} & A_{11} B_{12}+A_{12} B_{22}=0 \Rightarrow\left[B_{22}=A_{22}^{-1}\right] \Rightarrow \\
& A_{11} B_{12}+A_{12} A_{22}^{-1}=0 \Rightarrow A_{11} B_{12}=-A_{12} A_{22}^{-1} \Rightarrow \\
& B_{12}=-A_{11}^{-1} A_{12} A_{22}^{-1}
\end{array}
$$

Finally,

$$
B=\left(\begin{array}{c|c}
A_{11}^{-1} & -A_{11}^{-1} A_{12} A_{22}^{-1} \\
\hline 0 & A_{22}^{-1}
\end{array}\right)
$$

## Partitioned matrices

## Example

Computational Tomography (CT) with multiple rows gives a non-block structure for the system matrix that forces the problem to be solved in 3D. However, with a single row detector, the system matrix has a block structure so that the problem can be solved as a series of 2D problems strongly accelerating the process (on the other side the redundancy introduced by multiple row offers better resolution and robustness to noise).


## Exercises

## Exercises

From Lay (3rd ed.), Chapter 2, Section 4:

- 2.4.15
- 2.4.16
- 2.4.18
- 2.4.19


## Outline

(3) Matrix algebra

- Matrix operations (a)
- Inverse of a matrix (b)
- Elementary matrices (b)
- An algorithm to invert matrices (b)
- Characterization of invertible matrices (c)
- Invertible linear transformations (c)
- Partitioned matrices (c)
- LU factorization (d)
- An application to computer graphics and image processing (d)
- Subspaces of $\mathbb{R}^{n}(\mathrm{e})$
- Dimension and rank (e)


## LU factorization

## Example

Let us presume that we have a collection of equation systems

$$
\begin{aligned}
& A \mathbf{x}=\mathbf{b}_{1} \\
& A \mathbf{x}=\mathbf{b}_{2}
\end{aligned}
$$

and $A$ is not invertible, which could be an efficient way of solving all of them together? Factorize $A$ as $A=L U$ (see below) and solve the equation system in two steps. In fact the method is so efficient it is even used to solve a single equation system.

## LU factorization

## Definition 8.1 (LU factorization)

Let $A \in \mathcal{M}_{m \times n}$ that can be reduced to a reduced echelon form without row permutations. We can factorize $A$ as $A=L U$, where $L$ is an invertible, lower triangular matrix (with 1s in the main diagonal) of size $m \times m$ and $U$ is an upper triangular matrix of size $m \times n$.
MATLAB: $[L, U]=l u(A)$

## Example

Let $A \in \mathcal{M}_{4 \times 5}$. LU factorization will produce two matrices $L$ and $U$ may be of the following structure

$$
A=L U=\left(\begin{array}{llll}
1 & 0 & 0 & 0 \\
0 & 1 & 0 & 0 \\
0 & 0 & 1 & 0 \\
0 & 0 & 0 & 1
\end{array}\right)\left(\begin{array}{lllll}
\diamond & 0 & 0 & 0 & 0 \\
0 & \diamond & 0 & 0 & 0 \\
0 & 0 & 0 & \diamond & 0 \\
0 & 0 & 0 & 0 & 0
\end{array}\right)
$$

## LU factorization

Solving a linear equation system using the LU decomposition
Consider the equation system $A \mathbf{x}=\mathbf{b}$, and assume we have decomposed $A$ as $A=L U$. Then, we can solve the equation system in two steps:

$$
A \mathbf{x}=\mathbf{b} \Rightarrow(L U) \mathbf{x}=L(U \mathbf{x})=\mathbf{b} \Rightarrow\left\{\begin{array}{l}
L \mathbf{y}=\mathbf{b} \\
U \mathbf{x}=\mathbf{y}
\end{array}\right.
$$



FIGURE 2 Factorization of the mapping $\mathbf{x} \mapsto A \mathbf{x}$.

## LU factorization

## Example

Consider

$$
\left.A=\left(\begin{array}{rrrr}
3 & -7 & -2 & 2 \\
-3 & 5 & 1 & 0 \\
6 & -4 & 0 & -5 \\
-9 & 5 & -5 & 12
\end{array}\right)=\left(\begin{array}{rrrr}
1 & 0 & 0 & 0 \\
-1 & 1 & 0 & 0 \\
2 & -5 & 1 & 0 \\
-3 & 8 & 3 & 1
\end{array}\right)\left(\begin{array}{rrr}
3 & -7 & -2 \\
0 & -2 & -1 \\
0 & 0 & -1 \\
0 & 0 & 0
\end{array}\right)-1\right)
$$

and $\mathbf{b}=(-9,5,7,11)$. We first solve $L \mathbf{y}=\mathbf{b}$

$$
\left(\begin{array}{rrrl|r}
1 & 0 & 0 & 0 & -9 \\
-1 & 1 & 0 & 0 & 5 \\
2 & -5 & 1 & 0 & 7 \\
-3 & 8 & 3 & 1 & 11
\end{array}\right) \sim\left(\begin{array}{llll|r}
1 & 0 & 0 & 0 & -9 \\
0 & 1 & 0 & 0 & -4 \\
0 & 0 & 1 & 0 & 5 \\
0 & 0 & 0 & 1 & 1
\end{array}\right)
$$

and now we solve $U \mathbf{x}=\mathbf{y}$

$$
\left(\begin{array}{rrrr|r}
3 & -7 & -2 & 2 & -9 \\
0 & -2 & -1 & 2 & -4 \\
0 & 0 & -1 & 1 & 5 \\
0 & 0 & 0 & -1 & 1
\end{array}\right) \sim\left(\begin{array}{llll|r}
1 & 0 & 0 & 0 & 3 \\
0 & 1 & 0 & 0 & 4 \\
0 & 0 & 1 & 0 & -6 \\
0 & 0 & 0 & 1 & -1
\end{array}\right)
$$

The trick is that, thanks to the triangular structure, solving these two equation systems is rather fast.

## An algorithm to simple LU factorizations

## Algorithm

Let us assume that $A$ is row-equivalent to $U$ only using row replacement only with the rows above the replaced row. Then, there must be a sequence of elementary matrices such that

$$
A \sim U \Rightarrow E_{p} \ldots E_{2} E_{1} A=U \Rightarrow A=\left(E_{p} \ldots E_{2} E_{1}\right)^{-1} U
$$

By inspection, we note that $L=\left(E_{p} \ldots E_{2} E_{1}\right)^{-1}$.
In the previous algorithm we are making using of the following theorem:

## Theorem 8.1

(1) The product of two lower triangular matrices is lower triangular.
(2) The inverse of a lower triangular matrix is lower triangular.

## An algorithm to simple LU factorizations

## Example

$$
\begin{aligned}
& \mathbf{r}_{2} \leftarrow \mathbf{r}_{2}-\frac{1}{2} \mathbf{r}_{1} \quad E_{1}=\left(\begin{array}{ccc}
1 & 0 & 0 \\
-\frac{1}{2} & 1 & 0 \\
0 & 0 & 1
\end{array}\right) \quad\left(\begin{array}{lll}
2 & 1 & 0 \\
1 & 2 & 1 \\
0 & 1 & 2
\end{array}\right) \\
& \mathbf{r}_{3} \leftarrow \mathbf{r}_{3}-\frac{2}{3} \mathbf{r}_{2} \quad E_{2}=\left(\begin{array}{ccc}
1 & 0 & 0 \\
0 & 1 & 0 \\
0 & -\frac{2}{3} & 1
\end{array}\right) \quad U=\left(\begin{array}{lll}
2 & 1 & 0 \\
0 & \frac{3}{2} & 1 \\
0 & 1 & 2
\end{array}\right)
\end{aligned}
$$

Now, we calculate $L$ as

$$
L=\left(E_{2} E_{1}\right)^{-1}=E_{1}^{-1} E_{2}^{-1}=\left(\begin{array}{ccc}
1 & 0 & 0 \\
\frac{1}{2} & 1 & 0 \\
0 & 0 & 1
\end{array}\right)\left(\begin{array}{ccc}
1 & 0 & 0 \\
0 & 1 & 0 \\
0 & \frac{2}{3} & 1
\end{array}\right)=\left(\begin{array}{ccc}
1 & 0 & 0 \\
\frac{1}{2} & 1 & 0 \\
0 & \frac{2}{3} & 1
\end{array}\right)
$$

## An algorithm to simple LU factorizations

## Example

$$
\begin{aligned}
& \mathbf{r}_{2} \leftarrow \mathbf{r}_{2}-\frac{1}{2} \mathbf{r}_{1} \quad E_{1}=\left(\begin{array}{ccc}
1 & 0 & 0 \\
-\frac{1}{2} & 1 & 0 \\
0 & 0 & 1
\end{array}\right) \quad\left(\begin{array}{lll}
2 & 1 & 0 \\
1 & 2 & 1 \\
0 & 1 & 2
\end{array}\right) \\
& \mathbf{r}_{3} \leftarrow \mathbf{r}_{3}-\frac{2}{3} \mathbf{r}_{2} \quad E_{2}=\left(\begin{array}{ccc}
1 & 0 & 0 \\
0 & 1 & 0 \\
0 & -\frac{2}{3} & 1
\end{array}\right) \quad U=\left(\begin{array}{lll}
2 & 1 & 0 \\
0 & \frac{3}{2} & 1 \\
0 & 1 & 2
\end{array}\right)
\end{aligned}
$$

Now, we calculate $L$ as

$$
L=\left(E_{2} E_{1}\right)^{-1}=E_{1}^{-1} E_{2}^{-1}=\left(\begin{array}{ccc}
1 & 0 & 0 \\
\frac{1}{2} & 1 & 0 \\
0 & 0 & 1
\end{array}\right)\left(\begin{array}{ccc}
1 & 0 & 0 \\
0 & 1 & 0 \\
0 & \frac{2}{3} & 1
\end{array}\right)=\left(\begin{array}{ccc}
1 & 0 & 0 \\
\frac{1}{2} & 1 & 0 \\
0 & \frac{2}{3} & 1
\end{array}\right)
$$

## LDU factorization

## Example (continued)

Note that the $L$ and $U$ matrices found so far are assymetric in the sense that $L$ has 1 s in its main diagonal, but $U$ has not. We can extract the elements in the main diagonal of $U$ to a separate matrix $D$ by simply dividing the corresponding row of $U$ by that element:

$$
\begin{aligned}
& \begin{aligned}
A & =L U=\left(\begin{array}{lll}
1 & 0 & 0 \\
\frac{1}{2} & 1 & 0 \\
0 & \frac{2}{3} & 1
\end{array}\right)\left(\begin{array}{lll}
2 & 1 & 0 \\
0 & \frac{3}{2} & 1 \\
0 & 0 & \frac{4}{3}
\end{array}\right) \\
& =L D U=\left(\begin{array}{lll}
1 & 0 & 0 \\
\frac{1}{2} & 1 & 0 \\
0 & \frac{2}{3} & 1
\end{array}\right)\left(\begin{array}{lll}
1 & 0 & 0 \\
0 & \frac{3}{2} & 0 \\
0 & 0 & \frac{4}{3}
\end{array}\right)\left(\begin{array}{lll}
1 & \frac{1}{2} & 0 \\
0 & 1 & \frac{2}{3} \\
0 & 0 & 1
\end{array}\right)
\end{aligned} \\
& \text { diagonal matrix. } \\
& \text { where } D \text { is always a }
\end{aligned}
$$

## Other factorization examples

## Other factorizations

There are many other possibilities to factorize a matrix $A \in \mathcal{M}_{m \times n}$. See http://en.wikipedia.org/wiki/Matrix_decomposition. Among the most important are:

QR: $A=Q R$ where $Q \in \mathcal{M}_{m \times m}$ is orthogonal $\left(Q^{t} Q=D\right)$ and $R \in \mathcal{M}_{m \times n}$ is upper triangular.
SVD: $A=U D V^{t}$ where $U \in \mathcal{M}_{m \times m}$ is unitary $\left(U^{t} U=I_{m}\right), D \in \mathcal{M}_{m \times n}$ is diagonal, and $V \in \mathcal{M}_{n \times n}$ is also unitary ( $V^{t} V=I_{n}$ ).
Spectral: $A=P D P^{-1}$ (only for square matrices) where $P \in \mathcal{M}_{n \times n}$ and $D \in \mathcal{M}_{n \times n}$ is diagonal.

## Exercises

Exercises
From Lay (3rd ed.), Chapter 2, Section 5:

- 2.5.9
- 2.5.Practice problem


## Outline

(3) Matrix algebra

- Matrix operations (a)
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## An application to computer graphics and image processing

## Example

In vectorial graphics, graphics are described as a set of connected points (whose coordinates are known).


We may produce "italic" fonts by shearing the standard coordinates $T(\mathbf{x})=A \mathbf{x}$ where $A=$ $\left(\begin{array}{cc}1 & 0.25 \\ 0 & 1\end{array}\right)$.


## An application to computer graphics and image processing

## Example

Coordinate translations can be expressed as $T(\mathbf{x})=\mathbf{x}+\mathbf{x}_{0}$. But this is not a linear transformation:

$$
\begin{aligned}
T(\mathbf{u}) & =\mathbf{u}+\mathbf{x}_{0} \\
T(\mathbf{v}) & =\mathbf{v}+\mathbf{x}_{0} \\
T(\mathbf{u}+\mathbf{v}) & =\mathbf{u}+\mathbf{v}+\mathbf{x}_{0} \\
T(\mathbf{u})+T(\mathbf{v}) & =\left(\mathbf{u}+\mathbf{x}_{0}\right)+\left(\mathbf{v}+\mathbf{x}_{0}\right)=\mathbf{u}+\mathbf{v}+2 \mathbf{x}_{0} \\
T(\mathbf{u}+\mathbf{v}) & \neq T(\mathbf{u})+T(\mathbf{v})
\end{aligned}
$$

We can solve this problem with homogeneous coordinates.

## An application to computer graphics and image processing

## Definition 9.1 (Homogeneous coordinates)

Given a point with coordinates $\mathbf{x}$ we can construct its homogeneous coordinates as

$$
\tilde{\mathbf{x}}=\binom{h \mathbf{x}}{h}
$$

Or in other words, given the homogeneous coordinates $\tilde{\mathbf{u}}=\binom{\mathbf{u}}{h}$, they represent the point at $\frac{\mathrm{u}}{h}$. It is customary to use $h=1$ (but it is not compulsory, and in certain applications it is better to use other h's).

## Example

The 2D point $(1,2)$ can be represented in homogeneous coordinates as $(1,2,1)$, as $(2,4,2)$ and, even, as $(-2,-4,-2)$. They all represent the same point.

## An application to computer graphics and image processing

## Example

Now, coordinate translations in homogeneous coordinates is a linear transformation. For instance, in 2D:

$$
T(\tilde{\mathbf{x}})=A \tilde{\mathbf{x}}=\left(\begin{array}{ccc}
1 & 0 & \Delta x \\
0 & 1 & \Delta y \\
0 & 0 & 1
\end{array}\right)\left(\begin{array}{l}
x \\
y \\
1
\end{array}\right)=\left(\begin{array}{c}
x+\Delta x \\
y+\Delta y \\
1
\end{array}\right)
$$

## 2D transformations in homogeneous coordinates

In general, any 2D transformation of the form $T(\mathbf{x})=A \mathbf{x}$ can be represented in homogeneous coordinates as

$$
T(\tilde{\mathbf{x}})=\left(\begin{array}{ll}
A & \mathbf{0} \\
\mathbf{0} & 1
\end{array}\right) \tilde{\mathbf{x}}
$$

## An application to computer graphics and image processing

## Example

An application in 3D graphics:
http://www. youtube.com/watch?v=EsNmiiKlRXQ
Example
Let's say we want to
(1)

Rotate a point $30^{\circ}$ about the $Y$ axis.

(2) then, translate by $(-6,4,5)$

## An application to computer graphics and image processing

## Example (continued)

We need to use the transformation $T(\tilde{\mathbf{x}})=\tilde{A} \tilde{\mathbf{x}}$ with

$$
\tilde{A}=\left(\begin{array}{cccc}
1 & 0 & 0 & -6 \\
0 & 1 & 0 & 4 \\
0 & 0 & 1 & 5 \\
0 & 0 & 0 & 1
\end{array}\right)\left(\begin{array}{cccc}
\cos \left(30^{\circ}\right) & 0 & \sin \left(30^{\circ}\right) & 0 \\
0 & 1 & 0 & 0 \\
-\sin \left(30^{\circ}\right) & 0 & \cos \left(30^{\circ}\right) & 0 \\
0 & 0 & 0 & 1
\end{array}\right)
$$

and

$$
\tilde{x}=\left(\begin{array}{l}
x \\
y \\
z \\
1
\end{array}\right)
$$

## An application to computer graphics and image processing

## Example

Let's say we want to produce perspective projections. Let's imagine that the screen is on the $X Y$ plane and the viewer's eye is at $(0,0, d)$ (the distance to the screen is $d$ ). Any object between the viewer and the screen is projected onto the screen as in the figure below


By similar triangles we have

$$
\tan \alpha=\frac{x^{*}}{d}=\frac{x}{d-z} \Rightarrow x^{*}=\frac{x}{1-\frac{z}{d}}
$$

## An application to computer graphics and image processing

## Example (continued)

Similarly, $y^{*}=\frac{y}{1-\frac{z}{d}}$. Using homogeneous coordinates we want that $(x, y, z, 1)$ maps onto $\left(\frac{x}{1-\frac{z}{d}}, \frac{y}{1-\frac{z}{d}}, 0,1\right)$, or what is the same $\left(x, y, 0,1-\frac{z}{d}\right)$. We can achieve this with the perspective transformation:

$$
\tilde{P}=\left(\begin{array}{cccc}
1 & 0 & 0 & 0 \\
0 & 1 & 0 & 0 \\
0 & 0 & 0 & 0 \\
0 & 0 & -\frac{1}{d} & 1
\end{array}\right)
$$

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 2, Section 7:

- 2.7.2
- 2.7.3
- 2.7.10
- 2.7.12
- 2.7.22


## Outline

(3) Matrix algebra

- Matrix operations (a)
- Inverse of a matrix (b)
- Elementary matrices (b)
- An algorithm to invert matrices (b)
- Characterization of invertible matrices (c)
- Invertible linear transformations (c)
- Partitioned matrices (c)
- LU factorization (d)
- An application to computer graphics and image processing (d)
- Subspaces of $\mathbb{R}^{n}(\mathrm{e})$
- Dimension and rank (e)


## Subspace

## Definition 10.1 (Subspace of $\mathbb{R}^{n}$ )

$H \subseteq \mathbb{R}^{n}$ is a subspace of $\mathbb{R}^{n}$ if:
(1) $0 \in H$
(2) $\forall \mathbf{u}, \mathbf{v} \in H \quad \mathbf{u}+\mathbf{v} \in H$ (H is closed under vector addition)
( $\forall \mathbf{u} \in H \forall r \in \mathbb{R} \quad r \mathbf{u} \in H$ (H is closed under multiplication by a scalar)

## Example: Special subspaces

The following two sets are subspaces of $\mathbb{R}^{n}$ :
(1) $H=\{0\}$
(2) $H=\mathbb{R}^{n}$

## Subspace

## Example: Plane

A plane is defined as
$H=\operatorname{Span}\left\{\mathbf{v}_{1}, \mathbf{v}_{2}\right\}=\left\{\mathbf{v} \in \mathbb{R}^{n} \mid \mathbf{v}=\lambda_{1} \mathbf{v}_{1}+\lambda_{2} \mathbf{v}_{2}\right\}$
This plane is a subspace of $\mathbb{R}^{3}$
Proof

(1) Proof $\mathbf{0} \in H$

If $\lambda_{1}=\lambda_{2}=0$, then $\mathbf{v}=\mathbf{0}$.
(2) Proof $\mathbf{u}+\mathbf{v} \in H$

$$
\begin{aligned}
\mathbf{u} \in H & \Rightarrow \mathbf{u}=\lambda_{1 u} \mathbf{v}_{1}+\lambda_{2 u} \mathbf{v}_{2} \\
\mathbf{v} \in H & \Rightarrow \mathbf{v}=\lambda_{1 v} \mathbf{v}_{1}+\lambda_{2 v} \mathbf{v}_{2} \\
\mathbf{u}+\mathbf{v} & =\left(\lambda_{1 u} \mathbf{v}_{1}+\lambda_{2 u} \mathbf{v}_{2}\right)+\left(\lambda_{1 v} \mathbf{v}_{1}+\lambda_{2 v} \mathbf{v}_{2}\right) \\
& =\left(\lambda_{1 u}+\lambda_{1 v}\right) \mathbf{v}_{1}+\left(\lambda_{2 u}+\lambda_{2 v}\right) \mathbf{v}_{2} \in H
\end{aligned}
$$

(3) Proof $r \mathbf{u} \in H$

$$
\begin{aligned}
\mathbf{u} \in H & \Rightarrow \mathbf{u}=\lambda_{1 u} \mathbf{v}_{1}+\lambda_{2 u} \mathbf{v}_{2} \\
r \mathbf{u} & =r\left(\lambda_{1 u} \mathbf{v}_{1}+\lambda_{2 u} \mathbf{v}_{2}\right) \\
& =r \lambda_{1 u} \mathbf{v}_{1}+r \lambda_{2 u} \mathbf{v}_{2} \in H
\end{aligned}
$$

## Subspace

## Example: Line not through the origin

A line (L) that does not pass through the origin is not a subspace, because
(1) $0 \notin L$
(2) If we take two points belonging to the line ( $\mathbf{u}$ and $\mathbf{v}$ ), $\mathbf{u}+\mathbf{v} \notin L$.
(3) If we take a point belonging to the line ( $\mathbf{w}$ ), $2 \mathbf{w} \notin L$.

$\mathbf{u}+\mathbf{v}$ is not on $L$

$2 w$ is not on $L$

## Subspace

Example: Line through the origin
Consider $\mathbf{v}_{1}$ and $\mathbf{v}_{2}=k \mathbf{v}_{1}$. Then,

$$
H=\operatorname{Span}\left\{\mathbf{v}_{1}, \mathbf{v}_{2}\right\}=\operatorname{Span}\left\{\mathbf{v}_{1}\right\}
$$

is a line. It is easy to prove that this line is a subspace of $\mathbb{R}^{n}$.


$$
\mathbf{v}_{1} \neq \mathbf{0}, \mathbf{v}_{2}=k \mathbf{v}_{1} .
$$

## Column space

Definition 10.2 (Column space of a matrix)
Let $A \in \mathcal{M}_{m \times n}$. Let $\mathbf{a}_{i} \in \mathbb{R}^{m}$ be the columns of $A$. The column space of $A$ is defined as

$$
\operatorname{Col}\{A\}=\operatorname{Span}\left\{\mathbf{a}_{1}, \mathbf{a}_{2}, \ldots, \mathbf{a}_{n}\right\} \subseteq \mathbb{R}^{m}
$$

## Theorem 10.1

$\operatorname{Col}\{A\}$ is a subspace of $\mathbb{R}^{m}$.

## Column space

## Example

Let $A=\left(\begin{array}{rrr}1 & -3 & -4 \\ -4 & 6 & -2 \\ -3 & 7 & 6\end{array}\right)$ and $\mathbf{b}=\left(\begin{array}{r}3 \\ 3 \\ -4\end{array}\right)$.
Determine if $\mathbf{b}$ belongs to $\operatorname{Col}\{A\}$.
Solution
If $\mathbf{b} \in \operatorname{Col}\{A\}$ there must be some coefficients $x_{1}, x_{2}$ and $x_{3}$ such that

$$
\mathbf{b}=x_{1} \mathbf{a}_{1}+x_{2} \mathbf{a}_{2}+x_{3} \mathbf{a}_{3}
$$

To find these coefficients we simply have to solve the equation system $A \mathbf{x}=\mathbf{b}$.

$$
\left(\begin{array}{rrr|r}
1 & -3 & -4 & 3 \\
-4 & 6 & -2 & 3 \\
-3 & 7 & 6 & -4
\end{array}\right) \sim\left(\begin{array}{rrr|r}
1 & -3 & -4 & 3 \\
0 & -6 & -18 & 15 \\
0 & 0 & 0 & 0
\end{array}\right)
$$

In fact, there are infinite solutions to the equation system and, consequently, $\mathbf{b} \in \operatorname{Col}\{A\}$.

## Null space

## Definition 10.3 (Null space of a matrix)

Let $A \in \mathcal{M}_{m \times n}$. The null space of $A$ is defined as

$$
\operatorname{Nul}\{A\}=\left\{\mathbf{v} \in \mathbb{R}^{n} \mid A \mathbf{v}=\mathbf{0}\right\}
$$

## Theorem 10.2

$\operatorname{Nul}\{A\}$ is a subspace of $\mathbb{R}^{n}$.
Proof
(1) Proof $\mathbf{0} \in \operatorname{Nul}\{A\}$
$A \mathbf{0}=\mathbf{0} \Rightarrow \mathbf{0} \in \operatorname{Nul}\{A\}$ (q.e.d.)
(2) Proof $\mathbf{u}+\mathbf{v} \in \operatorname{Nul}\{A\}$
$\mathbf{u} \in \operatorname{Nul}\{A\} \Rightarrow A \mathbf{u}=\mathbf{0}$
$\mathbf{v} \in \operatorname{Nul}\{A\} \Rightarrow A \mathbf{v}=\mathbf{0}$
$A(\mathbf{u}+\mathbf{v})=A \mathbf{u}+A \mathbf{v}=\mathbf{0}+\mathbf{0}=\mathbf{0} \Rightarrow \mathbf{u}+\mathbf{v} \in \operatorname{Nul}\{A\}$ (q.e.d.)
(3) Proof ru $\in \operatorname{Nul}\{A\}$
$\mathbf{u} \in \operatorname{Nul}\{A\} \Rightarrow A \mathbf{u}=\mathbf{0}$
$A(r \mathbf{u})=r A \mathbf{u}=r \mathbf{0}=\mathbf{0} \Rightarrow r \mathbf{u} \in \operatorname{Nul}\{A\}$ (q.e.d.)

## Basis of a subspace

## Definition 10.4 (Basis of a subspace)

Let $H \subseteq \mathbb{R}^{n}$. The set of vectors $B$ is a basis of $H$ if:
(1) All vectors in $B$ are linearly independent
(2) $H=\operatorname{Span}\{B\}$

## Standard basis of $\mathbb{R}^{n}$

Let be the vectors

$$
\mathbf{e}_{1}=\left(\begin{array}{c}
1 \\
0 \\
0 \\
\ldots \\
0
\end{array}\right) \quad \mathbf{e}_{2}=\left(\begin{array}{c}
0 \\
1 \\
0 \\
\ldots \\
0
\end{array}\right) \quad \mathbf{e}_{3}=\left(\begin{array}{c}
0 \\
0 \\
1 \\
\ldots \\
0
\end{array}\right) \quad \ldots \quad \mathbf{e}_{n}=\left(\begin{array}{c}
0 \\
0 \\
0 \\
\ldots \\
1
\end{array}\right)
$$

The set $B=\left\{\mathbf{e}_{1}, \mathbf{e}_{2}, \ldots, \mathbf{e}_{n}\right\}$ is the standard basis of $\mathbb{R}^{n}$.

## Basis of a subspace

## Example

Find a basis for the null space of $A=\left(\begin{array}{rrrrr}-3 & 6 & -1 & 1 & -7 \\ 1 & -2 & 2 & 3 & -1 \\ 2 & -4 & 5 & 8 & -4\end{array}\right)$.
Solution
The null space of $A$ are all those vectors satisfying $A \mathbf{x}=\mathbf{0}$.

$$
(A \mid 0) \sim\left(\begin{array}{rrrrr|r}
1 & -2 & 0 & -1 & 3 & 0 \\
0 & 0 & 1 & 2 & -2 & 0 \\
0 & 0 & 0 & 0 & 0 & 0
\end{array}\right)
$$

So the solution of the equation system is $\left.\begin{array}{c}x_{1}=2 x_{2}+x_{4}-3 x_{5} \\ x_{3}=-2 x_{4}+2 x_{5}\end{array}\right\} \Rightarrow$

$$
\mathbf{x}=\left(\begin{array}{c}
2 x_{2}+x_{4}-3 x_{5} \\
x_{2} \\
-2 x_{4}+2 x_{5} \\
x_{4} \\
x_{5}
\end{array}\right)=x_{2}\left(\begin{array}{l}
2 \\
1 \\
0 \\
0 \\
0
\end{array}\right)+x_{4}\left(\begin{array}{r}
1 \\
0 \\
-2 \\
1 \\
0
\end{array}\right)+x_{5}\left(\begin{array}{r}
-3 \\
0 \\
2 \\
0 \\
1
\end{array}\right)
$$

## Null space and equation systems

## Example (continued)

The set $B=\{(2,1,0,0,0),(1,0,-2,1,0),(-3,0,2,0,1)\}$ is a basis of $\operatorname{Nul}\{A\}$. By construction, we have chosen them to be linearly independent.

## Example: Null space and equation systems

Consider $A=\left(\begin{array}{lll}1 & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 0 & 0\end{array}\right)$

- $\left\{\mathbf{e}_{3}\right\}$ is a basis for $\operatorname{Nul}\{A\}$
- Consider $\mathbf{b}=(7,3,0)$. The general solution of $A \mathbf{x}=\mathbf{b}$ is of the form

$$
\mathbf{x}=\mathbf{x}_{0}+\mathbf{x}_{N u l}
$$

where $\mathbf{x}_{0}$ is a solution of $A \mathbf{x}=\mathbf{b}$ that does not belong to $\operatorname{Nul}\{A\}$ and $\mathbf{x}_{N u l}$ belongs to $\operatorname{Nul}\{A\}$. In this particular case,

$$
\mathbf{x}=(7,3,0)+x_{3} \mathbf{e}_{3}
$$

## Null space and equation systems

## Example: Null space and equation systems (continued)

Let us prove that the general solution is actually a solution of $A \mathbf{x}=\mathbf{b}$

$$
A \mathbf{x}=A\left(\mathbf{x}_{0}+\mathbf{x}_{N u l}\right)=A \mathbf{x}_{0}+A \mathbf{x}_{N u l}=\mathbf{b}+\mathbf{0}=\mathbf{b}
$$

Intuititively we can say that the null space is the set of all solutions for which we have no measurements. The equation system only impose some constraints on those coefficients for which we have measurements. This is a problem in real situations as shown in the following slide.

## Null space and equation systems

In this example, the authors describe how the exact location of a tooth fracture is uncertain (Fig. C) due to the artifacts introduced by the null space of the tomographic problem.


Mora, M. A.; Mol, A.; Tyndall, D. A., Rivera, E. M. In vitro assessment of local computed tomography for the detection of longitudinal tooth fractures.
Oral Surg Oral Med Oral Pathol Oral Radiol Endod, 2007, 103, 825-829.

## Basis of a subspace

## Example

Find a basis for the column space of $B=\left(\begin{array}{rrrrr}1 & 0 & -3 & 5 & 0 \\ 0 & 1 & 2 & -1 & 0 \\ 0 & 0 & 0 & 0 & 1 \\ 0 & 0 & 0 & 0 & 0\end{array}\right)$.

## Solution

From the columns with non-pivot positions of matrix $B$ we learn that

$$
\begin{gathered}
\mathbf{b}_{3}=-3 \mathbf{b}_{1}+2 \mathbf{b}_{2} \\
\mathbf{b}_{4}=5 \mathbf{b}_{1}-\mathbf{b}_{2}
\end{gathered}
$$

Then,

$$
\left.\begin{array}{rl}
\operatorname{Col}\{B\} & =\left\{\mathbf{v} \in \mathbb{R}^{4} \mid \mathbf{v}=x_{1} \mathbf{b}_{1}+x_{2} \mathbf{b}_{2}+x_{3} \mathbf{b}_{3}+x_{4} \mathbf{b}_{4}+x_{5} \mathbf{b}_{5}\right\} \\
& =\left\{\mathbf{v} \in \mathbb{R}^{4} \left\lvert\, \begin{array}{cc}
\mathbf{v}= & x_{1} \mathbf{b}_{1}+x_{2} \mathbf{b}_{2}+x_{3}\left(-3 \mathbf{b}_{1}+2 \mathbf{b}_{2}\right)+ \\
& x_{4}\left(5 \mathbf{b}_{1}-\mathbf{b}_{2}\right)+x_{5} \mathbf{b}_{5}
\end{array}\right.\right\} \\
& =\left\{\mathbf{v} \in \mathbb{R}^{4} \mid \mathbf{v}=x_{1}^{\prime} \mathbf{b}_{1}+x_{2}^{\prime} \mathbf{b}_{2}+x_{5} \mathbf{b}_{5}\right\}
\end{array}\right\}
$$

And, consequently, $\operatorname{Basis}\{\operatorname{Col}\{B\}\}=\left\{\mathbf{b}_{1}, \mathbf{b}_{2}, \mathbf{b}_{5}\right\}$

## Basis of a subspace

## Example

Find a basis for the column space of $A=\left(\begin{array}{rrrrr}1 & 3 & 3 & 2 & -9 \\ -2 & -2 & 2 & -8 & 2 \\ 2 & 3 & 0 & 7 & 1 \\ 3 & 4 & -1 & 11 & -8\end{array}\right)$.

## Solution

It turns out that $A \sim B$ ( $B$ in the previous example). Since row operations do not affect linear dependence relations among the columns of the matrix, we should have

$$
\begin{gathered}
\mathbf{a}_{3}=-3 \mathbf{a}_{1}+2 \mathbf{a}_{2} \\
\mathbf{a}_{4}=5 \mathbf{a}_{1}-\mathbf{a}_{2}
\end{gathered}
$$

and $\operatorname{Basis}\{\operatorname{Col}\{A\}\}=\left\{\mathbf{a}_{1}, \mathbf{a}_{2}, \mathbf{a}_{5}\right\}$

## Theorem 10.3

The pivot columns of $A$ form a basis of $\operatorname{Col}\{A\}\}$.

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 2, Section 1:

- 2.8.1
- 2.8.2
- 2.8.5


## Outline

(3) Matrix algebra

- Matrix operations (a)
- Inverse of a matrix (b)
- Elementary matrices (b)
- An algorithm to invert matrices (b)
- Characterization of invertible matrices (c)
- Invertible linear transformations (c)
- Partitioned matrices (c)
- LU factorization (d)
- An application to computer graphics and image processing (d)
- Subspaces of $\mathbb{R}^{n}(e)$
- Dimension and rank (e)


## Coordinate system

## Definition 11.1 (Coordinates of a vector in the basis $B$ )

Suppose $B=\left\{\mathbf{b}_{1}, \mathbf{b}_{2}, \ldots, \mathbf{b}_{p}\right\}$ is a basis for the subspace $H \subseteq \mathbb{R}^{n}$. For each $\mathbf{x} \in H$, the coordinates of $\mathbf{x}$ relative to the basis $B$ are the weights $c_{i}$ such that

$$
\mathbf{x}=c_{1} \mathbf{b}_{1}+c_{2} \mathbf{b}_{2}+\ldots+c_{p} \mathbf{b}_{p}
$$

The coordinates of x with respect to the basis $B$ is the vector in $\mathbb{R}^{p}$

$$
[\mathbf{x}]_{B}=\left(\begin{array}{c}
c_{1} \\
c_{2} \\
\ldots \\
c_{p}
\end{array}\right)
$$

## Coordinate system

## Example

Let $\mathbf{x}=(3,12,7), \mathbf{v}_{1}=(3,6,2), \mathbf{v}_{2}=(-1,0,1), B=\left\{\mathbf{v}_{1}, \mathbf{v}_{2}\right\}$.
(1) Show that $B$ is a linearly independent set
(2) Find the coordinates of $\mathbf{x}$ in the coordinate system $B$

## Solution

(1) We need to prove that the only solution of the equation system $c_{1} \mathbf{v}_{1}+c_{2} \mathbf{v}_{2}=\mathbf{0}$ is $c_{1}=c_{2}=0$.

$$
\left(\begin{array}{rr|r}
3 & -1 & 0 \\
6 & 0 & 0 \\
2 & 1 & 0
\end{array}\right) \sim\left(\begin{array}{ll|l}
1 & 0 & 0 \\
0 & 1 & 0 \\
0 & 0 & 0
\end{array}\right)
$$

And, therefore, the unique solution is $c_{1}=c_{2}=0$ (q.e.d.)
(2) We need to find $c_{1}$ and $c_{2}$ such that $c_{1} \mathbf{v}_{1}+c_{2} \mathbf{v}_{2}=\mathbf{x}$

$$
\begin{aligned}
& \left(\begin{array}{rr|r}
3 & -1 & 3 \\
6 & 0 & 12 \\
2 & 1 & 7
\end{array}\right) \sim\left(\begin{array}{rr|r}
1 & 0 & 2 \\
0 & 1 & 3 \\
0 & 0 & 0
\end{array}\right) \\
& \text { And, therefore, }[\mathbf{x}]_{B}=(2,3) .
\end{aligned}
$$

## Coordinate system

## Example (continued)

The following figure shows how $\mathbf{x}$ is equal to $2 \mathbf{v}_{1}+3 \mathbf{v}_{2}$


## Coordinate system

## Theorem 11.1

The coordinates of a given vector with respect to a given basis are unique. Proof
Let us assume they are not unique. Then, there must be two different sets of coordinates such that

$$
\begin{aligned}
& \mathbf{x}=c_{1} \mathbf{b}_{1}+c_{2} \mathbf{b}_{2}+\ldots+c_{p} \mathbf{b}_{p} \\
& \mathbf{x}=c_{1}^{\prime} \mathbf{b}_{1}+c_{2}^{\prime} \mathbf{b}_{2}+\ldots+c_{p}^{\prime} \mathbf{b}_{p}
\end{aligned}
$$

If we subtract both equations, we have

$$
\mathbf{0}=\left(c_{1}-c_{1}^{\prime}\right) \mathbf{b}_{1}+\left(c_{2}-c_{2}^{\prime}\right) \mathbf{b}_{2}+\ldots+\left(c_{p}-c_{p}^{\prime}\right) \mathbf{b}_{p}
$$

But because the basis is a linearly independent set of vectors, it must be

$$
\begin{aligned}
& c_{1}-c_{1}^{\prime}=0 \Rightarrow c_{1}=c_{1}^{\prime} \\
& c_{2}-c_{2}^{\prime}=0 \Rightarrow c_{2}=c_{2}^{\prime} \\
& c_{p}-c_{p}^{\prime}=0 \Rightarrow c_{p}=c_{p}^{\prime}
\end{aligned}
$$

This is a contradiction with the hypothesis that there were two different sets of coordinates, and therefore, the coordinates of the vector $\mathbf{x}$ must be unique.

## Subspace dimension

## Isomorphism to $\mathbb{R}^{p}$

For any given subspace $H$ and its corresponding basis $B$, the mapping

$$
\begin{aligned}
T: H & \rightarrow \mathbb{R}^{p} \\
\mathbf{x} & \rightarrow[\mathbf{x}]_{B}
\end{aligned}
$$

is a linear, injective transformation that makes $H$ to behave as $\mathbb{R}^{p}$.

## Definition 11.2 (Dimension)

The dimension of a subspace $H(\operatorname{dim}\{H\})$ is the number of vectors of any of its basis.
The dimension of $\mathrm{H}=\{\mathbf{0}\}$ is 0 .

## Example (continued)

In our previous example in which $B=\left\{\mathbf{v}_{1}, \mathbf{v}_{2}\right\}$, the dimension is 2 , in fact $H$ behaves like a plane (see previous figure in the example).

## Rank of a matrix

## Definition 11.3 (Rank of a matrix)

The rank of a matrix $A$ is $\operatorname{rank}\{A\}=\operatorname{dim}\{\operatorname{Col}\{A\}\}$, that is, the dimension of the column space of the matrix.
MATLAB: $\operatorname{rank}(A)$

## Theorem 11.2

The rank of a matrix is the number of pivot columns it has.
Proof
Since the pivot columns form a basis of the column space of $A$, the number of pivot columns is the rank of the matrix.

## Example

$A=\left(\begin{array}{rrrrr}1 & 3 & 3 & 2 & -9 \\ -2 & -2 & 2 & -8 & 2 \\ 2 & 3 & 0 & 7 & 1 \\ 3 & 4 & -1 & 11 & -8\end{array}\right) \sim\left(\begin{array}{rrrrr}1 & 0 & -3 & 5 & 0 \\ 0 & 1 & 2 & -1 & 0 \\ 0 & 0 & 0 & 0 & 1 \\ 0 & 0 & 0 & 0 & 0\end{array}\right)$
Therefore, the rank of $A$ is 3 .

## Rank of a matrix

## Theorem 11.3 (Rank theorem)

If $A$ has $n$ columns, then

$$
\operatorname{Rank}\{A\}+\operatorname{dim}\{\operatorname{Nul}\{\mathrm{A}\}\}=n
$$

## Theorem 11.4 (Basis theorem)

Let $H$ be a subspace of dimension $p$. Any linearly independent set of $p$ vectors of $H$ is a basis of $H$. Any set of $p$ vectors that span $H$ is a basis of $H$.

## Characterization of invertible matrices (continued)

```
Theorem 11.5 (The invertible matrix theorem)
Let }A\in\mp@subsup{\mathcal{M}}{n\timesn}{}\mathrm{ . The following statements are equivalent (either they are all true or they are all false):
xiii. The columns of A form a basis of }\mp@subsup{\mathbb{R}}{}{n
xiv. }\operatorname{Col}{\textrm{A}}=\mp@subsup{\mathbb{R}}{}{n
xv. }\operatorname{dim}{\operatorname{Col}{A}}=
xvi. Rank}{A}=
xvii. Nul{A}}={\mathbf{0}
xviii. }\operatorname{dim}{\operatorname{Nul}{A}}=
```


## Characterization of invertible matrices

Proof $v \Rightarrow$ xiii<br>This is true by the basis theorem.<br>Proof xiii $\Rightarrow$ xiv<br>By the definition of basis.<br>Proof xiii $\Rightarrow x v$<br>By the definition of dimension.<br>Proof $x v \Rightarrow x v i$<br>By the definition of rank.<br>Proof xvi $\Rightarrow$ xviii<br>By the rank theorem.<br>Proof xvii $\Rightarrow$ iv<br>By the definition of null space.

## Exercises

```
Exercises
From Lay (3rd ed.), Chapter 2, Section 9:
- 2.9.1
- 2.9.3
- 2.9.9
- 2.9.19
- 2.9.27
```


## Outline

(3) Matrix algebra

- Matrix operations (a)
- Inverse of a matrix (b)
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- Dimension and rank (e)


# Chapter 4. Determinant of a matrix 

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## Outline

(4) Determinant of a matrix

- Introduction
- Properties of determinants
- Cramer's rule
- Matrix inversion
- Areas and volumes


## References


D. Lay. Linear algebra and its applications (3rd ed). Pearson (2006). Chapter 3.

## A little bit of history

The determinant of a matrix was first proposed by Seki Takakazu (1683) and Gottfried Leibniz (1693). Then Gabriel Cramer (1750) and Augustin Cauchy (1812) used them to solve problems in analytical geometry. Currently, they are not so much used in computational algebra, but they give important insights into the structure of a matrix.


## Applications

The determinant plays an important role in the analysis of Brownian motion. It was first described by Robert Brown in 1827 (looking at pollen grains in water). Albert Einstein published in 1905 a paper in which he explained brownian motion as the result of the hitting molecules to bigger particles. This served as a theoretical basis for a posterior experiment by Jean Perrin that confirmed the existence of atoms. Jean Perrin was Nobel Prize in 1926.


See video at https://www.youtube.com/watch?v=hy-clLi8gHg

## Outline

(4) Determinant of a matrix

- Introduction
- Properties of determinants
- Cramer's rule
- Matrix inversion
- Areas and volumes


## Cofactor

## Definition 1.1 (Cofactor)

The cofactor of the ij-th element of the matrix $A$ is

$$
C_{i j}=(-1)^{i+j}\left|A_{i j}\right|
$$

where $A_{i j}$ is the matrix that results after eliminating the $i$-th row and the $j$-th column from matrix $A$.

## Example

In the following example we calculate $A_{32}$

$$
\begin{aligned}
& {\left[\begin{array}{rrrr}
1 & -2 & 5 & 0 \\
2 & 0 & 4 & -1 \\
3 & 1 & 0 & 7 \\
0 & 4 & -2 & 0
\end{array}\right]} \\
& A_{32}=\left[\begin{array}{rrr}
1 & 5 & 0 \\
2 & 4 & -1 \\
0 & -2 & 0
\end{array}\right]
\end{aligned}
$$

## Determinant of a matrix

## Definition 1.2 (Determinant of a matrix)

The determinant of a square $n \times n$ matrix $A(|A|$ or $\operatorname{det}\{A\})$ is a mapping from $\mathcal{M}_{n \times n}$ onto $\mathbb{R}$ such that

$$
|A|=\left\{\begin{array}{cc}
A & n=1 \\
a_{11} C_{11}+a_{12} C_{12}+\ldots+a_{1 n} C_{1 n} & n \geq 2
\end{array}\right.
$$

where $a_{i j}$ is the $i j$-th element of matrix $A$.
$M A T L A B: \operatorname{det}(A)$

## Example

$$
\begin{aligned}
& \operatorname{det}\left\{\left(\begin{array}{rrr}
1 & 5 & 0 \\
2 & 4 & -1 \\
0 & -2 & 0
\end{array}\right)\right\}=1 \operatorname{det}\left\{\left(\begin{array}{rr}
4 & -1 \\
-2 & 0
\end{array}\right)\right\}-5 \operatorname{det}\left\{\left(\begin{array}{rr}
2 & -1 \\
0 & 0
\end{array}\right)\right\}+0 \operatorname{det}\left\{\left(\begin{array}{rr}
2 & 4 \\
0 & -2
\end{array}\right)\right\} \\
& \operatorname{det}\left\{\left(\begin{array}{rr}
4 & -1 \\
-2 & 0 \\
2 & -1 \\
0 & =1 \cdot(-2)-5 \cdot 0+0 \cdot(-4)=-2 \\
\operatorname{det}\left\{\left(\begin{array}{rr}
0 & 0
\end{array}\right)\right\} & =2 \operatorname{det}\{0\}-(-1) 4 \operatorname{det}\{-2\}=4 \cdot 0-(-1) \cdot(-2)=-2 \\
0 & 4 \\
2 & -2
\end{array}\right)\right\}=2 \operatorname{det}\{-2\}-4 \operatorname{det}\{0\}=2 \cdot 0-(-1) \cdot 0=0 \\
& \operatorname{det}\{0\}=2 \cdot(-2)-4 \cdot 0=-4
\end{aligned}
$$

## Determinant of a matrix

## Theorem 1.1

For $n \geq 2$, the determinant can be computed as a weighted sum of the cofactors along any row or column

$$
|A|=\sum_{j=1}^{n} a_{i j} C_{i j}=\sum_{i=1}^{n} a_{i j} C_{i j}
$$

## Example (continued)

$$
\begin{aligned}
\operatorname{det}\left\{\left(\begin{array}{rrr}
1 & 5 & 0 \\
2 & 4 & -1 \\
0 & -2 & 0
\end{array}\right)\right\} & =0 \cdot C_{13}-1 \cdot C_{23}+0 \cdot C_{33}=-2 \\
C_{23} & =(-1)^{2+3}\left|\begin{array}{rr}
1 & 5 \\
0 & -2
\end{array}\right|=-(1|(-2)|-5|(0)|)=2
\end{aligned}
$$

## Determinant of a matrix

Theorem 1.2 (Useful particular cases)

- For $n=2$,

$$
|A|=a_{11} a_{22}-a_{12} a_{21}
$$

- For $n=3$,

$$
|A|=a_{11} a_{22} a_{33}+a_{12} a_{23} a_{31}+a_{13} a_{21} a_{32}-a_{11} a_{23} a_{32}-a_{12} a_{21} a_{33}-a_{13} a_{22} a_{31}
$$



## Determinant of a matrix

## Theorem 1.3 (Useful particular cases (continued))

- For triangular matrices,

$$
|A|=\prod_{i=1}^{n} a_{i i}
$$

## Example

$$
\left|\begin{array}{rrrr}
1 & 4 & \frac{3}{5} & 2 \\
0 & 1 & 2 & -10 \\
0 & 0 & 1 & 12 \\
0 & 0 & 0 & 1
\end{array}\right|=1\left|\begin{array}{rrr}
1 & 2 & -10 \\
0 & 1 & 12 \\
0 & 0 & 1
\end{array}\right|=1 \cdot 1\left|\begin{array}{rr}
1 & 12 \\
0 & 1
\end{array}\right|=1 \cdot 1 \cdot 1|1|=1
$$

Computing the determinant requires $O(n!)$ operations if we do it through the cofactor expansion. There are much faster algorithms $\left(O\left(n^{3}\right)\right)$ that look for triangular matrices that have the same determinant as the original matrix and, then, they use this theorem that makes a much faster calculation.

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 3, Section 1:

- 3.1.42
- 3.1.43 (with computer; MATLAB: $A=r a n d(4))$
- 3.1.44 (with computer)
- 3.1.45 (with computer)
- 3.1.46 (with computer)


## Outline

(4) Determinant of a matrix

- Introduction
- Properties of determinants
- Cramer's rule
- Matrix inversion
- Areas and volumes


## Properties of determinants

$$
\begin{aligned}
& \text { Theorem } 2.1 \text { (Determinant of the multiplication) } \\
& \operatorname{det}\{A B\}=\operatorname{det}\{A\} \operatorname{det}\{B\} \\
& \operatorname{det}\{k A\}=k^{n} \operatorname{det}\{A\}
\end{aligned}
$$

Note: In general, $\operatorname{det}\{A+B\} \neq \operatorname{det}\{A\}+\operatorname{det}\{B\}$

## Theorem 2.2 (Determinant of row operations)

(1) If a multiple of one row of a matrix $A$ is added to another row to obtain a matrix $B$, then $\operatorname{det}\{B\}=\operatorname{det}\{A\}$.
(2) If two rows of a matrix $A$ are interchanged to obtain a matrix $B$, then $\operatorname{det}\{B\}=-\operatorname{det}\{A\}$.
(3) If a row of a matrix $A$ is multiplied by $k$ to obtain a matrix $B$, then $\operatorname{det}\{B\}=k \operatorname{det}\{A\}$.

## Properties of determinants

## Example

Consider the following transformations that are of the form $B=E A$
(1) $B=\left(\begin{array}{llll}1 & 0 & 0 & 0 \\ k & 1 & 0 & 0 \\ 0 & 0 & 1 & 0 \\ 0 & 0 & 0 & 1\end{array}\right) A \Rightarrow|B|=|E||A|=1|A|$
(2) $B=\left(\begin{array}{llll}0 & 1 & 0 & 0 \\ 1 & 0 & 0 & 0 \\ 0 & 0 & 1 & 0 \\ 0 & 0 & 0 & 1\end{array}\right) A \Rightarrow|B|=|E||A|=-1|A|$

- $B=\left(\begin{array}{llll}k & 0 & 0 & 0 \\ 0 & 1 & 0 & 0 \\ 0 & 0 & 1 & 0 \\ 0 & 0 & 0 & 1\end{array}\right) A \Rightarrow|B|=|E||A|=k|A|$


## Properties of determinants

## Example

## Properties of determinants

## Theorem 2.3

$A$ is invertible iff $|A| \neq 0$. In that case, $\left|A^{-1}\right|=|A|^{-1}$.
Corollary
If $|A|=0$, then the columns of $A$ are not linearly independent.

## Theorem 2.4

For any matrix $A \in \mathcal{M}_{n \times n}$, it is verified that $|A|=\left|A^{T}\right|$.

## Corollary

The effect of column operations on the determinant is the same as the effect of row operations.

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 3, Section 2:

- 3.2.14
- 3.2.15
- 3.2 .18
- 3.2.19
- 3.2.24
- 3.2.31
- 3.2.32
- 3.2.33
- 3.2.45 (computer)


## Outline

(4) Determinant of a matrix

- Introduction
- Properties of determinants
- Cramer's rule
- Matrix inversion
- Areas and volumes


## Cramer's rule

Cramer's rule is useful for a theoretical comprehension of what the determinant is and its properties, but it is not so useful for computational calculations.

## Theorem 3.1 (Cramer's rule)

Let $A \in \mathcal{M}_{n \times n}$ be an invertible matrix. For every $\mathbf{b} \in \mathbb{R}^{n}$ the $i$-th entry of the unique solution $\mathbf{x}$ of $A \mathbf{x}=\mathbf{b}$ is

$$
x_{i}=\frac{\operatorname{det}\left\{A_{i}(\mathbf{b})\right\}}{\operatorname{det}\{A\}}
$$

where $A_{i}(\mathbf{b})$ is the $A$ matrix in which the $i$-th column has been substituted by $\mathbf{b}$, that is,

$$
A_{i}(\mathbf{b})=\left(\begin{array}{llllllll}
\mathbf{a}_{1} & \mathbf{a}_{2} & \ldots & \mathbf{a}_{i-1} & \mathbf{b} & \mathbf{a}_{i+1} & \ldots & \mathbf{a}_{n}
\end{array}\right)
$$

Proof
Let $\mathbf{e}_{i}(i=1,2, \ldots, n)$ be the columns of the identity matrix $I_{n}$. Consider the product

$$
\left.\begin{array}{rl}
A l_{i}(\mathbf{x}) & =\left(\begin{array}{lllllllll}
A \mathbf{e}_{1} & A \mathbf{e}_{2} & \ldots & A \mathbf{e}_{i-1} & A \mathbf{x} & A \mathbf{e}_{i+1} & \ldots & A \mathbf{e}_{n}
\end{array}\right)= \\
& =\left(\begin{array}{llllll}
\mathbf{a}_{1} & \mathbf{a}_{2} & \ldots & \mathbf{a}_{i-1} & \mathbf{b} & \mathbf{a}_{i+1}
\end{array} \ldots\right. \\
\mathbf{a}_{n}
\end{array}\right)=A_{i}(\mathbf{b}) .40
$$

## Cramer's rule

Now we take the determinant on both sides

$$
\left|A_{i}(\mathbf{b})\right|=\left|A I_{i}(\mathbf{x})\right|=|A|\left|I_{i}(\mathbf{x})\right|=|A| x_{i} \Rightarrow x_{i}=\frac{\left|A_{i}(\mathbf{b})\right|}{|A|}
$$

## Example

Consider the equation system $\left(\begin{array}{cc}3 s & -2 \\ -6 & s\end{array}\right)\binom{x_{1}}{x_{2}}=\binom{4}{1}$. Its solution is given by

$$
\begin{aligned}
& x_{1}=\frac{\left|\begin{array}{rr}
4 & -2 \\
1 & s
\end{array}\right|}{\left|\begin{array}{rr}
3 s & -2 \\
-6 & s
\end{array}\right|}=\frac{4 s+2}{3 s^{2}-12}=\frac{4\left(s+\frac{1}{2}\right)}{3(s-2)(s+2)} \\
& x_{2}=\frac{\left|\begin{array}{rr}
3 s & 4 \\
-6 & 1
\end{array}\right|}{\left|\begin{array}{rr}
3 s & -2 \\
-6 & s
\end{array}\right|}=\frac{3 s+24}{3 s^{2}-12}=\frac{s+8}{(s-2)(s+2)}
\end{aligned}
$$

## Outline

(4) Determinant of a matrix

- Introduction
- Properties of determinants
- Cramer's rule
- Matrix inversion
- Areas and volumes


## Matrix inversion

## Algorithm to invert a matrix

We know that the inverse is a matrix such that $A A^{-1}=I_{n}$. If we call $\mathbf{x}_{i}$ to the $i$-th column of $A^{-1}$, then we have

$$
A A^{-1}=A\left(\begin{array}{llll}
\mathbf{x}_{1} & \mathbf{x}_{2} & \ldots & \mathbf{x}_{n}
\end{array}\right)=\left(\begin{array}{llll}
\mathbf{e}_{1} & \mathbf{e}_{2} & \ldots & \mathbf{e}_{n}
\end{array}\right)
$$

i.e., we are solving simultaneously $n$ equation systems of the form $A \mathbf{x}_{j}=\mathbf{e}_{j}$. The $i$-th entry of these columns is

$$
x_{i j}=\frac{\left|A_{i}\left(e_{j}\right)\right|}{|A|}
$$

If we now calculate the determinant in the numerator by expanding by the $j$-th column, we have $\left|A_{i}\left(\mathbf{e}_{j}\right)\right|=(-1)^{i+j}\left|A_{j i}\right|$, where $A_{j i}$ is the submatrix that results after eliminating the $j$-th row and the $i$-th column (or, what is the same, the cofactor of the $j i$-th element).

$$
x_{i j}=\frac{(-1)^{i+j}\left|A_{j i}\right|}{|A|}=\frac{c_{j i}}{|A|}
$$

## Matrix inversion

## Definition 4.1 (Adjoint (adjugate, adjunta) of a matrix)

Let $A \in \mathcal{M}_{n \times n}$ be a square matrix. The adjoint of $A$ is another $n \times n$ matrix, denoted by $A^{*}$ such that

$$
A_{i j}^{*}=C_{i j}
$$

## Algorithm to invert a matrix (continued)

Finally we have

$$
A^{-1}=\frac{1}{|A|}\left(\begin{array}{cccc}
C_{11} & C_{21} & \ldots & C_{n 1} \\
C_{12} & C_{22} & \ldots & C_{n 2} \\
\ldots & \ldots & \ldots & \ldots \\
C_{1 n} & C_{2 n} & \ldots & C_{n n}
\end{array}\right)
$$

Watch out that the indexes of the cofactors are transposed with respect to the standard order. Consequently

$$
A^{-1}=\frac{1}{|A|}\left(A^{T}\right)^{*}
$$

## Matrix inversion

## Theorem 4.1

$$
\left(A^{T}\right)^{*}=\left(A^{*}\right)^{T}
$$

## Example

$$
\begin{aligned}
& A=\left(\begin{array}{lll}
2 & 1 & 0 \\
1 & 1 & 0 \\
0 & 0 & 1
\end{array}\right) \Rightarrow
\end{aligned}
$$

$$
\begin{aligned}
& A^{*}=\left(\begin{array}{ccc}
1 & -1 & 0 \\
-1 & 2 & 0 \\
0 & 0 & 1
\end{array}\right) \Rightarrow A^{-1}=\frac{1}{|A|}\left(A^{*}\right)^{T}=\left(\begin{array}{ccc}
1 & -1 & 0 \\
-1 & 2 & 0 \\
0 & 0 & 1
\end{array}\right)^{T}=\left(\begin{array}{ccc}
1 & -1 & 0 \\
-1 & 2 & 0 \\
0 & 0 & 1
\end{array}\right)
\end{aligned}
$$

## Outline

(4) Determinant of a matrix

- Introduction
- Properties of determinants
- Cramer's rule
- Matrix inversion
- Areas and volumes


## Areas and volumes

Theorem 5.1 (Area of a parallelogram, Volume of a parallelepiped)
If $A$ is a $2 \times 2$ matrix, then $|\operatorname{det}\{A\}|$ is the area of the parallelogram formed by the columns of $A$. If $A$ is a $3 \times 3$ matrix, then $|\operatorname{det}\{A\}|$ is the volume of the parallelepiped formed by the columns of $A$.

## Example

Let be the parallelogram $A B C D(A=(-2,-2), B=(0,3), C=(4,-1)$, $D=(6,4)$ ).


The area can be calculated as

$$
\begin{aligned}
& |\operatorname{det}(\mathbf{B}-\mathbf{A} \mathbf{C}-\mathbf{A})|= \\
& \left|\operatorname{det}\left(\binom{0}{3}-\binom{-2}{-2}\binom{4}{-1}-\binom{-2}{-2}\right)\right|= \\
& \operatorname{det}\left(\begin{array}{ll}
2 & 6 \\
5 & 1
\end{array}\right)|=|-28|=28
\end{aligned}
$$

## Areas and volumes

## Theorem 5.2 (Area after a linear transformation)

Consider the transformation $T(\mathbf{x})=A \mathbf{x}$. If $A \in \mathcal{M}_{2 \times 2}$ and $S$ is a parallelogram in $\mathbf{R}^{2}$, then

$$
\operatorname{Area}\{T(S)\}=|\operatorname{det} A| \operatorname{Area}\{S\}
$$

If $A \in \mathcal{M}_{3 \times 3}$ and $S$ is a parallelepiped in $\mathbf{R}^{3}$, then the volume of $T(S)$ is

$$
\text { Volume }\{T(S)\}=|\operatorname{det} A| \text { Volume }\{S\}
$$

Proof
Let's prove it for the 2 D case (the 3D one is analogous).
Consider the columns of $A, A=\left(\begin{array}{ll}\mathbf{a}_{1} & \mathbf{a}_{2}\end{array}\right)$. Without loss of generality we may consider $S$ to be at the origin with sides given by $\mathbf{b}_{1}$ and $\mathbf{b}_{2}$ :

$$
S=\left\{\mathbf{x} \in \mathbb{R}^{2} \mid \mathbf{x}=s_{1} \mathbf{b}_{1}+s_{2} \mathbf{b}_{2} \forall s_{1}, s_{2} \in[0,1]\right\}
$$

## Areas and volumes

The image of $S$ by $T$ is

$$
T(S)=\left\{\mathbf{y} \in \mathbb{R}^{2} \mid \mathbf{y}=A \mathbf{x}=s_{1} A \mathbf{b}_{1}+s_{2} A \mathbf{b}_{2} \forall s_{1}, s_{2} \in[0,1]\right\}
$$

which is another parallelogram. Therefore, the area of $T(S)$ is

$$
\begin{aligned}
\text { Area }\{T(S)\} & =\left\lvert\, \operatorname{det}\left(\begin{array}{ll}
A \mathbf{b}_{1} & \left.A \mathbf{b}_{2}\right)\left|=\left|\operatorname{det}\left\{A\left(\begin{array}{ll}
\mathbf{b}_{1} & \mathbf{b}_{2}
\end{array}\right)\right\}\right|=|\operatorname{det}\{A B\}|\right. \\
& =|\operatorname{det} A||\operatorname{det} B|=|\operatorname{det} A| \text { Area }\{S\}
\end{array}\right.\right.
\end{aligned}
$$

(q.e.d.)

## Areas and volumes

## Theorem 5.3

The previous theorem is valid for any closed region in $\mathbb{R}^{2}$ or $\mathbb{R}^{3}$ with finite area or volume.
Proof (hint)
We only need to divide the region into very small (infinitely small) parallelograms (or parallelepipeds) and apply the previous theorem to each one of the pieces.


FIGURE 7 Approximating $T(R)$ by a union of parallelograms.

## Areas and volumes

## Example

Suppose that the unit disk defined as

$$
D=\left\{\mathbf{u} \in \mathbb{R}^{2} \mid u_{1}^{2}+u_{2}^{2} \leq 1\right\}
$$

is transformed with the transformation

$$
T(\mathbf{u})=\left(\begin{array}{ll}
a & 0 \\
0 & b
\end{array}\right) \mathbf{u}
$$

to produce

$$
E \equiv T(D)=\left\{\mathbf{x} \in \mathbb{R}^{2} \left\lvert\, \mathbf{x}=\left(\begin{array}{ll}
a & 0 \\
0 & b
\end{array}\right) \mathbf{u}=\binom{a u_{1}}{b u_{2}}\right.\right\}
$$

Exploiting the facts that $x_{1}=a u_{1} \Rightarrow u_{1}=\frac{x_{1}}{a}, x_{2}=b u_{2} \Rightarrow u_{2}=\frac{x_{2}}{b}$ we may also characterize the transformed region as

$$
E=\left\{\mathbf{x} \in \mathbb{R}^{2} \left\lvert\,\left(\frac{x_{1}}{a}\right)^{2}+\left(\frac{x_{2}}{b}\right)^{2} \leq 1\right.\right\}
$$

that is a solid ellipse.

## Areas and volumes

## Example (continued)



$$
\begin{aligned}
\text { Area }\{E\} & =|\operatorname{det} A| \operatorname{Area}\{D\}=(a b)\left(\pi(1)^{2}\right) \\
& =\pi a b
\end{aligned}
$$

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 3, Section 3:

- 3.3.1
- 3.3.7
- 3.3.11
- 3.3.21
- 3.3.25
- 3.3.26
- 3.3.29
- 3.3.32


## Outline

4 Determinant of a matrix

- Introduction
- Properties of determinants
- Cramer's rule
- Matrix inversion
- Areas and volumes


# Chapter 5. Vector spaces 

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## Outline

(5) Vector spaces

- Definition (a)
- Vector subspace (a)
- Subspace spanned by a set of vectors (a)
- Null space and column space of a matrix (b)
- Kernel and range of a linear transformation (b)
- Linearly independent sets and bases (b)
- Bases for $\operatorname{Nul}\{A\}$ and $\operatorname{Col}\{A\}$ (c)
- Coordinate system (c)
- Dimension of a vector space (d)
- Rank of a matrix (d)
- Change of basis (d)


## References


D. Lay. Linear algebra and its applications (3rd ed). Pearson (2006). Chapter 4.

## A little bit of history

Vectors were first used about 1636 in 2D and 3D to describe geometrical operations by René Descartes and Pierre de Fermat. In 1857 the notation of vectors and matrices was unified by Arthur Cayley. Giuseppe Peano was the firsst to give the modern definition of vector space in 1888, and Henri Lebesgue (about 1900) applied this theory to describe functional spaces as vector spaces.


## Applications

It is difficult to think a mathematical tool with more applications than vector spaces. Thanks to them we may sum forces, control devices, model complex systems, denoise images, ... They underlie all these processes and it is thank to them that we can "nicely" operate with vectors. They are a mathemtical structure that generalizes many other useful structures.


## Outline

(5) Vector spaces

- Definition (a)
- Vector subspace (a)
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- Linearly independent sets and bases (b)
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- Rank of a matrix (d)
- Change of basis (d)


## Vector space

## Definition 1.1 (Vector space)

A vector space is a non-empty set, $V$, of objects (called vectors) in which we define two operations: the sum among vectors and the multiplication by a scalar (an element of any field, $\mathbb{K}$ ), and that $\forall \mathbf{u}, \mathbf{v}, \mathbf{w} \in V$ and $\forall c, d \in \mathbb{K}$ it is verified that
(1) $\mathbf{u}+\mathbf{v} \in V$
(2) $\mathbf{u}+\mathbf{v}=\mathbf{v}+\mathbf{u}$
(3) $(\mathbf{u}+\mathbf{v})+\mathbf{w}=\mathbf{u}+(\mathbf{v}+\mathbf{w})$
(3) $\exists \mathbf{0} \in V \mid \mathbf{u}+\mathbf{0}=\mathbf{u}$
(5) $\forall \mathbf{u} \in V \quad \exists!\mathbf{w} \in V \mid \mathbf{u}+\mathbf{w}=\mathbf{0}$ (we normally write $\mathbf{w}=-\mathbf{u}$ )
(0) $c \mathbf{v} \in V$
(3) $c(\mathbf{u}+\mathbf{v})=c \mathbf{u}+c \mathbf{v}$
(8) $(c+d) \mathbf{u}=c \mathbf{u}+d \mathbf{u}$
(9) $c(d \mathbf{u})=(c d) \mathbf{u}$
(10) $\mathbf{1 u}=\mathbf{u}$

## Vector space

Theorem 1.1 (Other properties)
(1) $0 \mathbf{u}=\mathbf{0}$
(3) $\mathrm{c} \mathbf{0}=\mathbf{0}$
(3) $-\mathbf{u}=(-1) \mathbf{u}$

Watch out that 0 and 1 refer respectively to the neutral elements of the sum and multiplication in the field $\mathbb{K}$. -1 is the opposite number in $\mathbb{K}$ of 1 with respect to the sum of scalars.

## Example: $\mathbb{R}^{n}$


$\mathbb{R}^{n}$ is a vector space of finite dimension for any $n$. As well as $\mathbb{C}^{n}$.

## Vector space

## Example: Force fields in Physics

Consider $V$ to be the set of all arrows (directed line segments) in 3D. Two arrows are regarded as equal if they have the same length and direction. Define the sum of arrows and the multiplication by a scalar as shown below:


FIGURE 3 The parallelogram rule.


## Vector space

## Example: Force fields in Physics (continued)

Here is an example of the application of some of the properties of vector spaces


FIGURE $2 \mathbf{u}+\mathbf{v}=\mathbf{v}+\mathbf{u}$.


FIGURE $3(\mathbf{u}+\mathbf{v})+\mathbf{w}=\mathbf{u}+(\mathbf{v}+\mathbf{w})$.

With a force field we may define at every point in 3D space, which is the force that is applied.

Conservative force field


## Vector space

## Example: Infinite sequences

Let $S$ be the set of all infinite sequences of numbers

$$
\mathbf{u}=\left(\ldots, u_{-2}, u_{-1}, u_{0}, u_{1}, u_{2}, \ldots\right)
$$

Define the sum among two vectors and the multiplication by a scalar as

$$
\begin{gathered}
\mathbf{u}+\mathbf{v}=\left(\ldots, u_{-2}+v_{-2}, u_{-1}+v_{-1}, u_{0}+v_{0}, u_{1}+v_{1}, u_{2}+v_{2}, \ldots\right) \\
c \mathbf{u}=\left(\ldots, c u_{-2}, c u_{-1}, c u_{0}, c u_{1}, c u_{2}, \ldots\right)
\end{gathered}
$$

Digital Signal Processing


## Vector space

## Example: Polynomials of degree $n\left(\mathbb{P}_{n}\right)$

Let $\mathbb{P}_{n}$ be the set of all polynomials of degree $n$

$$
u(x)=u_{0}+u_{1} x+u_{2} x^{2}+\ldots+u_{n} x^{n}
$$

Define the sum among two vectors and the multiplication by a scalar as

$$
\begin{aligned}
(u+v)(x)= & \left(u_{0}+v_{0}\right)+\left(u_{1}+v_{1}\right) x+\left(u_{2}+v_{2}\right) x^{2}+\ldots+\left(u_{n}+v_{n}\right) x^{n} \\
& (c u)(x)=c u_{0}+c u_{1} x+c u_{2} x^{2}+\ldots+c u_{n} x^{n}
\end{aligned}
$$

Legendre polynomials


## Vector space

## Example: Set of real functions defined in some domain

Let $\mathbb{F}$ be the set of all real valued functions defined in some domain $(f: D \rightarrow \mathbb{R})$ Define the sum among two vectors and the multiplication by a scalar as

$$
\begin{gathered}
(u+v)(x)=u(x)+v(x) \\
(c u)(x)=c u(x)
\end{gathered}
$$

Ex: $u(x)=3+x$
Ex: $v(x)=\sin x$
Ex: Zernike polynomials


## Outline

(5) Vector spaces

- Definition (a)
- Vector subspace (a)
- Subspace spanned by a set of vectors (a)
- Null space and column space of a matrix (b)
- Kernel and range of a linear transformation (b)
- Linearly independent sets and bases (b)
- Bases for $\operatorname{Nul}\{A\}$ and $\operatorname{Col}\{A\}$ (c)
- Coordinate system (c)
- Dimension of a vector space (d)
- Rank of a matrix (d)
- Change of basis (d)


## Vector subspace

Sometimes we don't need to deal with the whole vector space, but only a part of it. It would be nice if it also has the space properties.

## Definition 2.1 (Vector subspace)

Let $V$ be a vector space, and $H \subseteq V$ a part of it. $H$ is vector subspace iff
a) $\mathbf{0} \in H$
b) $\forall \mathbf{u}, \mathbf{v} \in H \quad \mathbf{u}+\mathbf{v} \in H$ (H is closed with respect to sum)
c) $\forall \mathbf{u} \in H, \forall c \in \mathbb{K} \quad c \mathbf{u} \in H$ ( $H$ is closed with respect to scalar multiplication)

## Example

$$
H=\{\mathbf{0}\} \text { is a subspace. }
$$

## Example

The vector space of polynomials (of any degree), $\mathbb{P} \in \mathbb{F}(\mathbb{R})$, is a vector subspace of the vector space of real valued functions defined over $\mathbb{R}(\mathbb{F}(\mathbb{R})=\{f: \mathbb{R} \rightarrow \mathbb{R}\})$.

## Vector subspace

## Example

$H=\mathbb{R}^{2}$ is not a subspace of $\mathbb{R}^{3}$ because $\mathbb{R}^{2} \not \subset \mathbb{R}^{3}$, for instance, the vector $\mathbf{u}=\binom{1}{2} \in \mathbb{R}^{2}$, but $\mathbf{u} \notin \mathbb{R}^{3}$.

## Example

$H=\mathbb{R}^{2} \times\{0\}$ is a subspace of $\mathbb{R}^{3}$ because all vectors of $H$ are of the form
$\mathbf{u}=\left(\begin{array}{c}x_{1} \\ x_{2} \\ 0\end{array}\right) \in \mathbb{R}^{3}$. It is obvious that $H$ "looks like" $\mathbb{R}^{2}$. This resemblance is mathematically called isomorphism.

## Example

Any plane in 3D passing through the origin is a subspace of $\mathbb{R}^{3}$. Any plane in 3D not passing through the origin is not a subspace of $\mathbb{R}^{3}$, because $\mathbf{0}$ does not belong to the plane.

## Vector subspace

Theorem 2.1
If $H$ is a vector subspace, then $H$ is a vector space.
Proof
a) $\Rightarrow 4$

$$
\begin{aligned}
& a \equiv \mathbf{0} \in H \\
& 4 \equiv \exists \mathbf{0} \in V \mid \mathbf{u}+\mathbf{0}=\mathbf{u}
\end{aligned}
$$

b) $\Rightarrow 1$
$b \equiv \forall \mathbf{u}, \mathbf{v} \in H \quad \mathbf{u}+\mathbf{v} \in H$
$1 \equiv \mathbf{u}+\mathbf{v} \in V$
Since $H \subset V$ and thanks to $b) \Rightarrow 2,3,7,8,9,10$

$$
2 \equiv \mathbf{u}+\mathbf{v}=\mathbf{v}+\mathbf{u}
$$

$$
3 \equiv(\mathbf{u}+\mathbf{v})+\mathbf{w}=\mathbf{u}+(\mathbf{v}+\mathbf{w})
$$

$$
7 \equiv c(\mathbf{u}+\mathbf{v})=c \mathbf{u}+c \mathbf{v}
$$

$$
8 \equiv(c+d) \mathbf{u}=c \mathbf{u}+d \mathbf{u}
$$

$$
9 \equiv c(d \mathbf{u})=(c d) \mathbf{u}
$$

$$
10 \equiv 1 \mathbf{u}=\mathbf{u}
$$

## Vector subspace

Proof (continued)
c) $\Rightarrow 6$
$c \equiv \forall \mathbf{u} \in H, \forall c \in \mathbb{K} \quad c \mathbf{u} \in H$
$6 \equiv c \mathbf{v} \in V$
Proof of 5
Since $H$ is a subset of $V$, we know that for every $\mathbf{u} \in H$ there exists
a unique $\mathbf{w} \in V \mid \mathbf{u}+\mathbf{w}=\mathbf{0}$. The problem is whether
or not $\mathbf{w}$ is in $H$. We also know that $\mathbf{w}=(-1) \mathbf{v}$, and
by $c), \mathbf{w} \in H$.
(q.e.d.)

## Outline

(5) Vector spaces

- Definition (a)
- Vector subspace (a)
- Subspace spanned by a set of vectors (a)
- Null space and column space of a matrix (b)
- Kernel and range of a linear transformation (b)
- Linearly independent sets and bases (b)
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- Change of basis (d)


## Subspace spanned by a set of vectors

## Example

Let $\mathbf{v}_{1}, \mathbf{v}_{2} \in V$ be two vectors of a vector space, $V$. The subset

$$
H=\operatorname{Span}\left\{\mathbf{v}_{1}, \mathbf{v}_{2}\right\}
$$

is a subspace of $V$.

## Proof

Any vector of $H$ is of the form $\mathbf{v}=\lambda_{1} \mathbf{v}_{1}+\lambda_{2} \mathbf{v}_{2}$ for some $\lambda_{1}, \lambda_{2} \in \mathbb{K}$.
Proof a) $\mathbf{0} \in H$
Simply by setting $\lambda_{1}=\lambda_{2}=0$, we get $\mathbf{0} \in H$
Proof b) $\mathbf{u}+\mathbf{v} \in H$
Let $\left.\mathbf{u}, \mathbf{v} \in H \Rightarrow \begin{array}{l}\mathbf{u}=\lambda_{1 u} \mathbf{v}_{1}+\lambda_{2 u} \mathbf{v}_{2} \\ \mathbf{v}=\lambda_{1 v} \mathbf{v}_{1}+\lambda_{2 v} \mathbf{v}_{2}\end{array}\right\} \Rightarrow$

$$
\begin{aligned}
\mathbf{u}+\mathbf{v} & =\left(\lambda_{1 u} \mathbf{v}_{1}+\lambda_{2 u} \mathbf{v}_{2}\right)+\left(\lambda_{1 v} \mathbf{v}_{1}+\lambda_{2 v} \mathbf{v}_{2}\right) \\
& =\left(\lambda_{1 u}+\lambda_{1 v}\right) \mathbf{v}_{1}+\left(\lambda_{2 u}+\lambda_{2 v}\right) \mathbf{v}_{2} \in H
\end{aligned}
$$

## Subspace spanned by a set of vectors

$$
\begin{aligned}
& \text { Proof } c) c \mathbf{u} \in H \\
& \text { Let } \mathbf{u} \in H \Rightarrow \\
& \qquad \mathbf{u}=\lambda_{1} \mathbf{v}_{1}+\lambda_{2} \mathbf{v}_{2} \Rightarrow c \mathbf{u}=c\left(\lambda_{u} \mathbf{v}_{1}+\lambda_{2} \mathbf{v}_{2}\right)=c \lambda_{u} \mathbf{v}_{1}+c \lambda_{2} \mathbf{v}_{2} \in H
\end{aligned}
$$

## Theorem 3.1

Let $\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{p} \in V$ be $p$ vectors of a vector space, $V$. The subset

$$
H=\operatorname{Span}\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{p}\right\}
$$

is a subspace of $V$.
Proof
Analogous to the previous example.

## Subspace spanned by a set of vectors

## Example

Consider the set of vectors $\mathbb{R}^{4} \supset H=\{(a-3 b, b-a, a, b) \forall a, b \in \mathbb{R}\}$. Is it a vector subspace?

## Solution

All vectors of $H$ can be written as

$$
H \ni \mathbf{u}=\left(\begin{array}{c}
a-3 b \\
b-a \\
a \\
b
\end{array}\right)=a\left(\begin{array}{c}
1 \\
-1 \\
1 \\
0
\end{array}\right)+b\left(\begin{array}{c}
-3 \\
1 \\
0 \\
1
\end{array}\right)
$$

Therefore, $H=\operatorname{Span}\{(1,-1,1,0),(-3,1,0,1)\}$ and by the previous theorem, it is a vector subspace.

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 4, Section 1:

- 4.1.1
- 4.1.4
- 4.1.5
- 4.1.6
- 4.1.19
- 4.1.32
- 4.1.37 (computer)


## Outline

(5) Vector spaces

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- Rank of a matrix (d)
- Change of basis (d)


## Null space of a matrix

## Example

Consider the matrix

$$
\left(\begin{array}{ccc}
1 & -3 & -2 \\
-5 & 9 & 1
\end{array}\right)
$$

The point $\mathbf{x}=(5,3,-2)$ has the property that $A \mathbf{x}=\mathbf{0}$.

## Definition 4.1 (Null space)

The null space of a matrix $A \in \mathcal{M}_{m \times n}$ is the set of vectors

$$
\operatorname{Nul}\{A\}=\left\{\mathbf{x} \in \mathbb{R}^{n} \mid A \mathbf{x}=\mathbf{0}\right\}
$$



## Null space of a matrix

## Example (continued)

$$
\left(\begin{array}{rrr|r}
1 & -3 & -2 & 0 \\
-5 & 9 & 1 & 0
\end{array}\right) \sim\left(\begin{array}{lll|l}
1 & 0 & \frac{5}{3} & 0 \\
0 & 1 & \frac{3}{2} & 0
\end{array}\right)
$$

Therefore

$$
\operatorname{Nul}\{A\}=\left\{\left(-\frac{5}{2} x_{3},-\frac{3}{2} x_{3}, x_{3}\right) \forall x_{3} \in \mathbb{R}\right\}
$$

The previous example $(\mathbf{x}=(5,3,-2))$ is the point we obtain for $x_{3}=-2$.

## Null space of a matrix

## Theorem 4.1

$\operatorname{Nul}\{A\}$ is a vector subspace of $\mathbb{R}^{n}$.
Proof
It is obvious that $\operatorname{Nul}\{A\} \subseteq \mathbb{R}^{n}$ because $A$ has $n$ columns
Proof a) $\mathbf{0} \in \operatorname{Nul}\{A\}$
$\overline{A 0}_{n}=\mathbf{0}_{m} \Rightarrow \mathbf{0}_{n} \in \operatorname{Nul}\{A\}$
$\underline{\text { Proof b) } \mathbf{u}+\mathbf{v} \in \operatorname{Nul}\{A\}}$
Let $\left.\mathbf{u}, \mathbf{v} \in \operatorname{Nul}\{A\} \Rightarrow \begin{array}{l}A \mathbf{u}=\mathbf{0} \\ A \mathbf{v}=\mathbf{0}\end{array}\right\} \Rightarrow$

$$
A(\mathbf{u}+\mathbf{v})=A \mathbf{u}+A \mathbf{v}=\mathbf{0}+\mathbf{0}=\mathbf{0} \Rightarrow \mathbf{u}+\mathbf{v} \in \operatorname{Nul}\{A\}
$$

Proof c) cu $\in \operatorname{Nul}\{A\}$
$\overline{\text { Let } \mathbf{u} \in H \Rightarrow}$

$$
A \mathbf{u}=\mathbf{0} \Rightarrow A(c \mathbf{u})=c(A \mathbf{u})=c \mathbf{0}=\mathbf{0} \Rightarrow c \mathbf{u} \in \operatorname{Nul}\{A\}
$$

## Null space of a matrix

## Example

Let $H=\left\{\begin{array}{l|l}(a, b, c, d) \in \mathbb{R}^{4} & \begin{array}{l}a-2 b+5 c=d \\ c-a=b\end{array}\end{array}\right\}$. Is $H$ a vector subspace of $\mathbb{R}^{4}$ ?
Solution
We may rewrite the conditions of belonging to $H$ as

$$
\begin{aligned}
& a-2 b+5 c=d \\
& c-a=b
\end{aligned} \Rightarrow\left(\begin{array}{rrrr}
1 & -2 & 5 & -1 \\
-1 & -1 & 1 & 0
\end{array}\right)\left(\begin{array}{l}
a \\
b \\
c \\
d
\end{array}\right)=\mathbf{0}
$$

and, thanks to the previous theorem, $H$ is a vector subspace of $\mathbb{R}^{4}$.

## Null space of a matrix

## Example (continued)

We can even provide a basis for $H$

$$
\left(\begin{array}{rrrr}
1 & -2 & 5 & -1 \\
-1 & -1 & 1 & 0
\end{array}\right) \sim\left(\begin{array}{rrrr}
1 & 0 & 1 & -1 \\
0 & 1 & 0 & 0
\end{array}\right)
$$

The solution of $A \mathbf{x}=\mathbf{0}$ are all points of the form

$$
\left(\begin{array}{l}
a \\
b \\
c \\
d
\end{array}\right)=\left(\begin{array}{c}
-c+d \\
0 \\
c \\
d
\end{array}\right)=c\left(\begin{array}{c}
-1 \\
0 \\
1 \\
0
\end{array}\right)+d\left(\begin{array}{l}
1 \\
0 \\
0 \\
1
\end{array}\right)
$$

Consequently $H=\operatorname{Span}\{(-1,0,1,0),(1,0,0,1)\}$.

## Column space of a matrix

## Definition 4.2 (Column space)

Let $A \in \mathcal{M}_{m \times n}$ a matrix and $\mathbf{a}_{i} \in \mathbb{R}^{m}(i=1,2, \ldots n)$ its columns. The column space of the matrix $A$ is defined as

$$
\operatorname{Col}\{A\}=\operatorname{Span}\left\{\mathbf{a}_{1}, \mathbf{a}_{2}, \ldots \mathbf{a}_{n}\right\}=\left\{\mathbf{b} \in \mathbb{R}^{m} \mid A \mathbf{x}=\mathbf{b} \text { for some } \mathbf{x} \in \mathbb{R}^{n}\right\}
$$

## Theorem 4.2

The column space of a matrix is a subspace of $\mathbb{R}^{m}$ Proof $\operatorname{Col}\{A\}$ is a set generated by a number of vectors and by Theorem 3.1 it is a subspace of $\mathbb{R}^{m}$.

## Column space of a matrix

## Example

Find a matrix $A$ such that $\operatorname{Col}\{A\}=\{(6 a-b, a+b,-7 a) \forall a, b \in \mathbb{R}\}$ Solution We can express the points in $\operatorname{Col}\{A\}$ as

$$
\operatorname{Col}\{A\} \ni \mathbf{x}=\left(\begin{array}{c}
6 a-b \\
a+b \\
-7 a
\end{array}\right)=a\left(\begin{array}{c}
6 \\
1 \\
-7
\end{array}\right)+b\left(\begin{array}{c}
-1 \\
1 \\
0
\end{array}\right)
$$

Therefore, $\operatorname{Col}\{A\}=\operatorname{Span}\{(6,1,-7),(-1,1,0)\}$. That is, these must be the two columns of $A$

$$
A=\left(\begin{array}{cc}
6 & -1 \\
1 & 1 \\
-7 & 0
\end{array}\right)
$$

## Comparison between the Null and the Column spaces

Contrast Between NuI $A$ and $\operatorname{Col} A$ for an $m \times n$ Matrix $A$

## $\operatorname{Nul} A \quad \operatorname{Col} A$

1. $\mathrm{Nul} A$ is a subspace of $\mathbb{R}^{n}$.
2. $\operatorname{Nul} A$ is implicitly defined; that is, you are given only a condition $(A \mathbf{x}=0)$ that vectors in $\mathrm{Nul} A$ must satisfy.
3. It takes time to find vectors in $\operatorname{Nul} A$. Row operations on $\left[\begin{array}{ll}A & 0\end{array}\right]$ are required.
4. There is no obvious relation between $\operatorname{Nul} A$ and the entries in $A$.
5. A typical vector $\mathbf{v}$ in $\mathrm{Nul} A$ has the property that $A \mathbf{v}=\mathbf{0}$.
6. Given a specific vector $\mathbf{v}$, it is easy to tell if $\mathbf{v}$ is in $\mathrm{Nul} A$. Just compute $A \mathbf{v}$.
7. $\operatorname{Nul} A=\{0\}$ if and only if the equation $A \mathbf{x}=0$ has only the trivial solution.
8. $\operatorname{Nul} A=\{0\}$ if and only if the linear transformation $\mathbf{x} \mapsto A \mathbf{x}$ is one-to-one.
9. $\operatorname{Col} A$ is a subspace of $\mathbb{R}^{m}$.
10. $\operatorname{Col} A$ is explicitly defined; that is, you are told how to build vectors in $\operatorname{Col} A$.
11. It is easy to find vectors in $\operatorname{Col} A$. The columns of $A$ are displayed; others are formed from them.
12. There is an obvious relation between $\operatorname{Col} A$ and the entries in $A$, since each column of $A$ is in $\mathrm{Col} A$.
13. A typical vector $\mathbf{v}$ in $\operatorname{Col} A$ has the property that the equation $A \mathbf{x}=\mathbf{v}$ is consistent.
14. Given a specific vector $\mathbf{v}$, it may take time to tell if $\mathbf{v}$ is in $\mathrm{Col} A$. Row operations on [ $\begin{array}{ll}A & \mathbf{v} \text { ] are required. }\end{array}$
15. $\operatorname{Col} A=\mathbb{R}^{m}$ if and only if the equation $A \mathbf{x}=\mathbf{b}$ has a solution for every $\mathbf{b}$ in $\mathbb{R}^{m}$.
16. $\operatorname{Col} A=\mathbb{R}^{m}$ if and only if the linear transformation $\mathbf{x} \mapsto A \mathbf{x}$ maps $\mathbb{R}^{n}$ onto $\mathbb{R}^{m}$.

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## Linear transformation

We have said that $T(\mathbf{x})=A \mathbf{x}$ is a linear transformation, but it is not the only one.

## Definition 5.1 (Linear transformation)

The transformation $T: V \rightarrow W$ between two vectors spaces $V$ and $W$ is a rule that for each vector $\mathbf{v} \in V$ assigns a unique vector $\mathbf{w}=T(\mathbf{v}) \in W$, such that
(1) $T\left(\mathbf{v}_{1}+\mathbf{v}_{2}\right)=T\left(\mathbf{v}_{1}\right)+T\left(\mathbf{v}_{2}\right) \quad \forall \mathbf{v}_{1}, \mathbf{v}_{2} \in V$
(2) $T(c \mathbf{v})=c T(\mathbf{v}) \quad \forall \mathbf{v} \in V, \forall c \in \mathbb{K}$

## Example

For a matrix $A \in \mathcal{M}_{m \times n}$, we have that

$$
\begin{aligned}
T: \mathbb{R}^{n} & \rightarrow \mathbb{R}^{m} \\
\mathbf{x} & \rightarrow A \mathbf{x}
\end{aligned}
$$

is a linear transformation (we can easily verify that $T$ meets the two required conditions).

## Linear transformation

## Example

Consider the space of all continuous, real-valued functions defined over $\mathbb{R}$ whose all derivatives are also continuous. We will refer to this space as $C^{\infty}(\mathbf{R})$. For instance, all polynomials belong to this space, as well as any sin, cos function. It can be proved that $C^{\infty}(\mathbf{R})$ is a vector space.
Consider the transformation that assigns to each function in $C^{\infty}(\mathbf{R})$ its derivative

$$
\begin{aligned}
D: C^{\infty}(\mathbf{R}) & \rightarrow C^{\infty}(\mathbf{R}) \\
f & \rightarrow D(f)
\end{aligned}
$$

is a linear transformation.
Proof
(1) $D(f+g)=D(f)+D(g)$
(2) $D(c f)=c D(f)$

## Kernel and range of transformation

## Definition 5.2 (Kernel (Núcleo))

The kernel of a transformation $T$ is the set of all vectors such that

$$
\operatorname{Ker}\{T\}=\{\mathbf{v} \in V \mid T(\mathbf{v})=\mathbf{0}\}
$$

## Definition 5.3 (Range (Imagen))

The range of a transformation $T$ is the set of all vectors such that

$$
\text { Range }\{T\}=\{\mathbf{w} \in W \mid \exists \mathbf{v} \in V T(\mathbf{v})=\mathbf{w}\}
$$



## Kernel and range of transformation

```
Example (continued)
\(\operatorname{Ker}\{T\}=\operatorname{Nul}\{A\}\)
\(\operatorname{Ker}\{D\}=\{f(x)=c\}\) because \(D(c)=0\)
```

Theorem 5.1
If $T(\mathbf{x})=A \mathbf{x}$, then

$$
\begin{aligned}
\operatorname{Ker}\{T\} & =\operatorname{Nul}\{A\} \\
\operatorname{Range}\{T\} & =\operatorname{Col}\{A\}
\end{aligned}
$$

## Exercises

```
Exercises
From Lay (3rd ed.), Chapter 4, Section 2:
    - 4.2.3
    - 4.2.9
    - 4.2.11
    - 4.2.30
    - 4.2.31
```


## Outline

(5) Vector spaces

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## Linear independence

## Definition 6.1 (Linear independence)

$A$ set of vectors $\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{p}\right\}$ is linearly independent iff the only solution to the equation

$$
c_{1} \mathbf{v}_{1}+c_{2} \mathbf{v}_{2}+\ldots+c_{p} \mathbf{v}_{p}=\mathbf{0}
$$

is the trivial solution $\left(c_{1}=c_{2}=\ldots=c_{p}=0\right)$. The set is linearly dependent if there exists another solution to the equation.

Watch out that we cannot simply put all vectors as columns of a matrix $A$ and solve $A \mathbf{c}=\mathbf{0}$ because this is only valid for vectors in $\mathbb{R}^{n}$, but it is not valid for any vector space.

## Linear independence

## Example

- $\left\{\mathbf{v}_{1}\right\}$ is linearly dependent if $\mathbf{v}_{1}=\mathbf{0}$.
- $\left\{\mathbf{v}_{1}, \mathbf{v}_{2}\right\}$ is linearly dependent if $\mathbf{v}_{2}=c \mathbf{v}_{1}$.
- $\left\{\mathbf{0}, \mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{p}\right\}$ is linearly dependent.


## Example

In the vector space of continuous functions over $\mathbb{R}, C(\mathbb{R})$, the vectors $f_{1}(x)=\sin x$ and $f_{2}(x)=\cos x$ are independent because

$$
f_{2}(x) \neq c f_{1}(x)
$$



## Linear independence

## Theorem 6.1

A set of vectors $\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{p}\right\}$, with $\mathbf{v}_{1} \neq \mathbf{0}$ is linearly dependent if any of the vectors $\mathbf{v}_{j}(j>1)$ is linearly dependent on the previous ones $\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{j-1}\right\}$.

## Example

In the vector space of polynomials, consider the vectors $p_{0}(x)=1, p_{1}(x)=x$, $p_{2}(x)=4-x$. The set $\left\{p_{0}(x), p_{1}(x), p_{2}(x)\right\}$ is linearly dependent because

$$
p_{2}(x)=4 p_{0}(x)-p_{1}(x) \Rightarrow p_{1}(x)-4 p_{0}(x)+p_{2}(x)=0
$$

## Linear independence

## Example

In the vector space of continuous functions, consider the vectors $f_{1}(x)=\sin (x) \cos (x)$ and $f_{2}(x)=\sin (2 x)$. The set $\left\{f_{1}(x), f_{2}(x)\right\}$ is linearly dependent because $f_{2}(x)=2 f_{1}(x)$

MATLAB:

```
x=[-pi:0.001:pi]
f1=sin(x).*cos(x);
f2=sin(2*x);
plot(x,f1,x,f2)
```



## Basis of a subspace

## Definition 6.2 (Basis of a subspace)

A set of vectors $B=\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{p}\right\}$ is a basis of the vector subspace $H$ iff
(1) $B$ is a linearly independent set of vectors
(2) $H=\operatorname{Span}\{B\}$

In other words, a basis is a non-redundant set of vectors that span H .

## Example

Let $A$ be an invertible matrix. By Theorem 5.1 and 11.5 of Chapter 3 (the invertible matrix theorem), we know that the columns of $A$ span $\mathbb{R}^{n}$ and that they are linearly independent. Consequently, the columns of $A$ are a basis of $\mathbb{R}^{n}$.

## Basis of a subspace

## Example

The standard basis of $\mathbb{R}^{n}$ are the columns of $I_{n}$

$$
\mathbf{e}_{1}=\left(\begin{array}{c}
1 \\
0 \\
\ldots \\
0
\end{array}\right) \quad \mathbf{e}_{2}=\left(\begin{array}{c}
0 \\
1 \\
\ldots \\
0
\end{array}\right) \quad \ldots \quad \mathbf{e}_{n}=\left(\begin{array}{c}
0 \\
0 \\
\ldots \\
1
\end{array}\right)
$$

## Example

Let $\mathbf{v}_{1}=(3,0,-6), \mathbf{v}_{2}=(-4,1,7), \mathbf{v}_{3}=(-2,1,5)$. Is $\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \mathbf{v}_{3}\right\}$ a basis of $\mathbb{R}^{3}$ ?

## Solution

This question is the same as whether $A$ is invertible with

$$
A=\left(\begin{array}{rrr}
3 & -4 & -2 \\
0 & 1 & 1 \\
-6 & 7 & 5
\end{array}\right) \Rightarrow|A|=6 \Rightarrow \exists A^{-1}
$$

Because $A$ is invertible, we have that $\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \mathbf{v}_{3}\right\}$ is a basis of $\mathbb{R}^{3}$.

## Basis of a subspace

## Example: DNA Structure

In 1953, Rosalind Franklin, James Watson and Francis Crick determined the 3D structure of DNA using data coming from X-ray diffraction of crystallized DNA. Watson and Crick received the Nobel prize in physiology and medicine in 1962 (Franklin died 1958).


## Basis of a subspace

## Example: DNA Structure (continued)

Three-dimensional crystals repeat a certain motif all over the space following a crystal lattice. The vectors that define the crystal lattice are a basis of $\mathbb{R}^{3}$


## Basis of a subspace

## Example

$B=\left\{1, x, x^{2}, x^{3}, \ldots\right\}$ is the standard basis of the vector space of polynomials $\mathbb{P}$.
Proof
(1) $B$ is linearly independent:

$$
\forall x \in \mathbb{R} \quad c_{0} 1+c_{1} x+c_{2} x^{2}+c_{3} x^{3}+\ldots=0 \Rightarrow c_{0}=c_{1}=c_{2}=\ldots=0
$$

The only way that a polynomial of degree whichever is 0 for all values of $x$ is that the coefficients of the polynomial are all 0 .
(2) $\mathbb{P}=\operatorname{Span}\{B\}$ :

It is obvious that any polynomial can be written as a linear combination of elements of $B$ (in fact, this is they way we normally do).

## Basis of a subspace

## Example

$H=\operatorname{Span}\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \mathbf{v}_{3}\right\}$ with $\mathbf{v}_{1}=(0,2,-1), \mathbf{v}_{2}=(2,2,0), \mathbf{v}_{3}=(6,16,-5)$. Find a basis of $H$

## Solution

All vectors in $H$ are of the form:

$$
H \ni \mathbf{x}=c_{1} \mathbf{v}_{1}+c_{2} \mathbf{v}_{2}+c_{3} \mathbf{v}_{3}
$$

We realize that $\mathbf{v}_{3}=5 \mathbf{v}_{1}+3 \mathbf{v}_{2}$, therefore, $\mathbf{v}_{3}$ is redundant:

$$
\begin{aligned}
H \ni \mathbf{x} & =c_{1} \mathbf{v}_{1}+c_{2} \mathbf{v}_{2}+c_{3}\left(5 \mathbf{v}_{1}+3 \mathbf{v}_{2}\right) \\
& =\left(c_{1}+5 c_{3}\right) \mathbf{v}_{1}+\left(c_{2}+3 c_{3}\right) \mathbf{v}_{2} \\
& =c_{1}^{\prime} \mathbf{v}_{1}+c_{2}^{\prime} \mathbf{v}_{2}
\end{aligned}
$$

It suffices to construct our basis with $\mathbf{v}_{1}$ and $\mathbf{v}_{2}$.

## Basis of a subspace

## Theorem 6.2 (Spanning set theorem (conjunto generador))

Let $S=\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{p}\right\}$ be a set of vectors and $H=\operatorname{Span}\{S\}$. Then,
(1) If $\mathbf{v}_{k}$ is a linear combination of the rest, then the set $S-\left\{\mathbf{v}_{k}\right\}$ still generates H.
(2) If $H \neq\{\mathbf{0}\}$, then some subset of $S$ is a basis of $H$.

Proof
(1) Assume that the linear combination that explains $\mathbf{v}_{k}$ is

$$
\mathbf{v}_{k}=a_{1} \mathbf{v}_{1}+\ldots+a_{k-1} \mathbf{v}_{k-1}+a_{k+1} \mathbf{v}_{k+1}+\ldots+a_{p} \mathbf{v}_{p}
$$

Consider any vector in $H$

$$
\begin{aligned}
\mathbf{x}= & c_{1} \mathbf{v}_{1}+c_{2} \mathbf{v}_{2}+\ldots+c_{p} \mathbf{v}_{p} \\
= & \left(c_{1}+a_{1}\right) \mathbf{v}_{1}+\ldots+\left(c_{k-1}+a_{k-1}\right) \mathbf{v}_{k-1}+ \\
& \left(c_{k+1}+a_{k+1}\right) \mathbf{v}_{k+1}+\ldots+\left(c_{p}+a_{p}\right) \mathbf{v}_{p}
\end{aligned}
$$

That is we can express $\mathbf{x}$ not using $\mathbf{v}_{k}$.
(2) Step 1: If $S$ is a linearly independent set, then $S$ is the basis of $H$. Step 2: If $S$ is not, using the previous point we can remove a vector to produce $S^{\prime}$ that still generates H (go to Step 1).

## Outline

(5) Vector spaces

- Definition (a)
- Vector subspace (a)
- Subspace spanned by a set of vectors (a)
- Null space and column space of a matrix (b)
- Kernel and range of a linear transformation (b)
- Linearly independent sets and bases (b)
- Bases for $\operatorname{Nul}\{A\}$ and $\operatorname{Col}\{A\}$ (c)
- Coordinate system (c)
- Dimension of a vector space (d)
- Rank of a matrix (d)
- Change of basis (d)


## Basis for $\operatorname{Nul}\{A\}$

## Example

Let $A=\left(\begin{array}{rrrrr}-3 & 6 & -1 & 1 & -7 \\ 1 & -2 & 2 & 3 & -1 \\ 2 & -4 & 5 & 8 & -4\end{array}\right)$
We solve the equation system $A \mathbf{x}=\mathbf{0}$ to find

$$
(A \mid \mathbf{0}) \sim\left(\begin{array}{rrrrr|r}
1 & -2 & 0 & -1 & 3 & 0 \\
0 & 0 & 1 & 2 & -2 & 0 \\
0 & 0 & 0 & 0 & 0 & 0
\end{array}\right)
$$

we have coloured the pivot columns from which learn

$$
\begin{aligned}
& x_{1}=2 x_{2}+x_{4}-3 x_{5} \\
& x_{3}=-2 x_{4}+2 x_{5}
\end{aligned} \Rightarrow \operatorname{Nul}\{A\} \ni \mathbf{x}=\left(\begin{array}{c}
2 x_{2}+x_{4}-3 x_{5} \\
x_{2} \\
-2 x_{4}+2 x_{5} \\
x_{4} \\
x_{5}
\end{array}\right)
$$

## Basis for $\operatorname{Nul}\{A\}$

## Example (continued)

$$
\operatorname{Nul}\{A\} \ni \mathbf{x}=\left(\begin{array}{c}
2 x_{2}+x_{4}-3 x_{5} \\
x_{2} \\
-2 x_{4}+2 x_{5} \\
x_{4} \\
x_{5}
\end{array}\right)=x_{2}\left(\begin{array}{l}
2 \\
1 \\
0 \\
0 \\
0
\end{array}\right)+x_{4}\left(\begin{array}{c}
1 \\
0 \\
-2 \\
1 \\
0
\end{array}\right)+x_{5}\left(\begin{array}{c}
-3 \\
0 \\
2 \\
0 \\
1
\end{array}\right)
$$

Finally the basis for $\operatorname{Nul}\{A\}$ is

$$
\operatorname{Nul}\{A\}=\operatorname{Span}\left\{\left(\begin{array}{l}
2 \\
1 \\
0 \\
0 \\
0
\end{array}\right),\left(\begin{array}{c}
1 \\
0 \\
-2 \\
1 \\
0
\end{array}\right),\left(\begin{array}{c}
-3 \\
0 \\
2 \\
0 \\
1
\end{array}\right)\right\}
$$

## Basis for $\operatorname{Col}\{A\}$

## Example

Consider $A$ as in the previous example. We had

$$
A \sim\left(\begin{array}{rrrrr}
1 & -2 & 0 & -1 & 3 \\
0 & 0 & 1 & 2 & -2 \\
0 & 0 & 0 & 0 & 0
\end{array}\right)=B
$$

Let's call this latter matrix $B$. Non-pivot columns of $B$ can be written as a linear combination of the pivot columns:

$$
\begin{aligned}
\mathbf{b}_{2} & =-2 \mathbf{b}_{1} \\
\mathbf{b}_{4} & =-\mathbf{b}_{1}+2 \mathbf{b}_{3} \\
\mathbf{b}_{5} & =3 \mathbf{b}_{1}-2 \mathbf{b}_{3}
\end{aligned}
$$

## Basis for $\operatorname{Col}\{A\}$

## Example (continued)

Since row operations do not change the linear dependences among matrix columns, we can derive the same relationships for matrix $A$

$$
\begin{aligned}
& \mathbf{a}_{2}=-2 \mathbf{a}_{1} \\
& \mathbf{a}_{4}=-\mathbf{a}_{1}+2 \mathbf{a}_{3} \\
& \mathbf{a}_{5}=3 \mathbf{a}_{1}-2 \mathbf{a}_{3}
\end{aligned}
$$

Finally, the basis of $\operatorname{Col}\{A\}$ is $\left\{\mathbf{a}_{1}, \mathbf{a}_{3}\right\}$.

$$
\operatorname{Col}\{A\}=\operatorname{Span}\left\{\mathbf{a}_{1}, \mathbf{a}_{3}\right\}=\operatorname{Span}\left\{\left(\begin{array}{c}
-3 \\
1 \\
2
\end{array}\right),\left(\begin{array}{c}
-1 \\
2 \\
5
\end{array}\right)\right\}
$$

## Basis for $\operatorname{Col}\{A\}$

## Theorem 7.1

The pivot columns of $A$ constitute a basis for $\operatorname{Col}\{A\}$.
Proof
Let $B$ the reduced echelon form of $A$.
(1) The pivot columns of $B$ form a linearly independent set because none of its elements can be expressed as a linear combination of the elements before each one of them.
(2) The dependence relationships among columns are not affected by row operations. Therefore, the corresponding pivot columns of $A$ are also linearly independent and, consequently, a basis of $\operatorname{Col}\{A\}$.

## Two views of a basis

## As small as possible, as large as possible

(1) The Spanning Set Theorem states that the basis is as small as possible as long as it spans the required subspace.
(2) The basis has the maximum amount of vectors spanning the required subspace. If we add one more, the new set is not linearly independent.

## Example

- $\{(1,0,0),(2,3,0)\}$ is a set of 2 linearly independent vectors. But it cannot span $\mathbb{R}^{3}$ because for this we need 3 vectors.
- $\{(1,0,0),(2,3,0),(4,5,6)\}$ is a set of 3 linearly independent vectors that spans $\mathbb{R}^{3}$, so it is a basis of $\mathbb{R}^{3}$.
- $\{(1,0,0),(2,3,0),(4,5,6),(7,8,9)\}$ is a set of 4 linearly dependent vectors that spans $\mathbb{R}^{3}$, so it cannot be a basis.


## Exercises

## Exercises

From Lay (3rd ed.), Chapter 4, Section 3:

- 4.3.1
- 4.3.2
- 4.3.8
- 4.3.12
- 4.3.24
- 4.3.31
- 4.3.32
- 4.3.33
- 4.3.37 (computer)


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## Coordinate system

An important reason to assign a basis to a vector space $V$ is that it makes $V$ to "behave" as $\mathbb{R}^{n}$ through, what is called, a coordinate system.

## Theorem 8.1 (The unique representation theorem)

Let $B=\left\{\mathbf{b}_{1}, \mathbf{b}_{2}, \ldots, \mathbf{b}_{n}\right\}$ a basis of the vector space $V$, and consider any vector $\mathbf{v} \in V$. There exists a unique set of scalars such that

$$
\mathbf{v}=c_{1} \mathbf{b}_{1}+c_{2} \mathbf{b}_{2}+\ldots+c_{n} \mathbf{b}_{n}
$$

## Proof

Let assume that there exists another set of scalars such that

$$
\mathbf{v}=c_{1}^{\prime} \mathbf{b}_{1}+c_{2}^{\prime} \mathbf{b}_{2}+\ldots+c_{n}^{\prime} \mathbf{b}_{n}
$$

Subtracting both equations we have

$$
\mathbf{0}=\left(c_{1}-c_{1}^{\prime}\right) \mathbf{b}_{1}+\left(c_{2}-c_{2}^{\prime}\right) \mathbf{b}_{2}+\ldots+\left(c_{n}-c_{n}^{\prime}\right) \mathbf{b}_{n}
$$

But since the vectors $\mathbf{b}_{i}$ form a basis and are linearly independent, it must be

$$
\left(c_{1}-c_{1}^{\prime}\right)=\left(c_{2}-c_{2}^{\prime}\right)=\left(c_{n}-c_{n}^{\prime}\right)=0
$$

## Coordinate system

Proof (continued)
Finally, $c_{1}=c_{1}^{\prime}, c_{2}=c_{2}^{\prime}, \ldots, c_{n}=c_{n}^{\prime}$ which is a contradiction with the hypothesis that there were two different sets of scalars representing the vector. Consequently, the set of scalars must be unique.

## Definition 8.1 (Coordinates)

Let $B=\left\{\mathbf{b}_{1}, \mathbf{b}_{2}, \ldots, \mathbf{b}_{n}\right\}$ a basis of the vector space $V$, and consider any vector $\mathbf{v} \in V$. The coordinates of $\mathbf{v}$ in $B$ are the $c_{i}$ coefficients such that

$$
\mathbf{v}=c_{1} \mathbf{b}_{1}+c_{2} \mathbf{b}_{2}+\ldots+c_{n} \mathbf{b}_{n} \Rightarrow[\mathbf{v}]_{B}=\left(\begin{array}{c}
c_{1} \\
c_{2} \\
\ldots \\
c_{n}
\end{array}\right)
$$

The transformation $T: V \rightarrow \mathbb{R}^{n}$ such that $T(\mathbf{x})=[\mathbf{x}]_{B}$ is called the coordinate mapping.

## Coordinate system

## Example

Let $B=\{(1,0),(1,2)\}$ be a basis of $\mathbb{R}^{2}$ and $[\mathbf{x}]_{B}=(-2,3)$, then

$$
\mathbf{x}=-2 \mathbf{b}_{1}+3 \mathbf{b}_{2}=-2\binom{1}{0}+3\binom{1}{2}=\binom{1}{6}
$$

In fact $(1,6)$ are the coordinates of $\mathbf{x}$ in the standard basis $\left\{\mathbf{e}_{1}, \mathbf{e}_{2}\right\}$

$$
\mathbf{x}=1 \mathbf{e}_{1}+6 \mathbf{e}_{2}=1\binom{1}{0}+6\binom{0}{1}=\binom{1}{6}
$$

That is, the point $\mathbf{x}$ does not change, but depending on the coordinate system employed, we "see" it with different coordinates.

## Coordinate system

## Example (continued)



FIGURE 1 Standard graph paper.


FIGURE $2 \mathcal{B}$-graph paper.

## Coordinate system

## Example: X-ray diffraction

In ths figure we see how a X-ray diffraction pattern of a crystal is "indexed".


## Coordinates in $\mathbb{R}^{n}$

If we have a point $\mathbf{x}$ in $\mathbb{R}$ we can easily find its coordinates in any basis, as in the following example.

## Example

Let $\mathbf{x}=(4,5)$ and the basis $B=\{(2,1),(-1,1)\}$. We need to find $c_{1}$ and $c_{2}$ such that

$$
\mathbf{x}=c_{1} \mathbf{b}_{1}+c_{2} \mathbf{b}_{2} \Rightarrow\binom{4}{5}=c_{1}\binom{2}{1}+c_{2}\binom{-1}{1}=\left(\begin{array}{cc}
2 & -1 \\
1 & 1
\end{array}\right)\binom{c_{1}}{c_{2}}
$$

From which we can easily derive that $c_{1}=3$ and $c_{2}=2$.


FIGURE 4
The $\mathcal{B}$-coordinate vector of $\mathbf{x}$ is $(3,2)$.

## Change of basis

## Change from the standard basis to an arbitrary basis

Note that the previous equation system is of the form

$$
\mathbf{x}=P_{B}[\mathbf{x}]_{B}
$$

where $P_{B}$ is called the change-of-coordinates matrix and its columns are the vectors of the basis $B$ (consequently, it is invertible). We find the coordinates of the vector $\mathbf{x}$ in the basis $B$ as

$$
[\mathbf{x}]_{B}=P_{B}^{-1} \mathbf{x}
$$

## Change between two arbitrary bases

Let's say we know the coordinates of a point in some basis, $B_{1}$, and we want to know its coordinates in some other basis, $B_{2}$. We may use

$$
\mathbf{x}=P_{B_{1}}[\mathbf{x}]_{B_{1}}=P_{B_{2}}[\mathbf{x}]_{B_{2}} \Rightarrow[\mathbf{x}]_{B_{2}}=P_{B_{2}}^{-1} P_{B_{1}}[\mathbf{x}]_{B_{1}}
$$

## Coordinate mapping

Theorem 8.2 (The coordinate mapping is an isomorphism between $V$ and $\mathbb{R}^{n}$ )
The coordinate mapping is a bijective, linear transformation.


FIGURE 5 The coordinate mapping from $V$ onto $\mathbb{R}^{n}$.

## Corollary

Since the coordinate mapping is a linear transformation it extends to linear combinations

$$
\left[a_{1} \mathbf{u}_{1}+a_{2} \mathbf{u}_{2}+\ldots+a_{p} \mathbf{u}_{p}\right]_{B}=a_{1}\left[\mathbf{u}_{1}\right]_{B}+a_{2}\left[\mathbf{u}_{2}\right]_{B}+\ldots+a_{p}\left[\mathbf{u}_{p}\right]_{B}
$$

## Coordinate mapping

## Consequences

Any operation in $V$ can be performed in $\mathbb{R}^{n}$ and then go back to $V$. For spaces of functions, this opens a new door to analyze functions (signals, images, ...) in $\mathbb{R}^{n}$ using the appropriate basis: Fourier transform, wavelet transform, Discrete Cosine Transform, ...


## Coordinate mapping

## Example

Consider the space of polynomials of degree $2, \mathbb{P}_{2}$. any polynomial in this space is of the form

$$
p(t)=a_{0}+a_{1} t+a_{2} t^{2}
$$

If we choose the standard basis in $\mathbb{P}_{2}$ that is

$$
B=\left\{1, t, t^{2}\right\}
$$

Then, we have the coordinate mapping

$$
T(p(t))=[p]_{B}=\left(\begin{array}{l}
a_{0} \\
a_{1} \\
a_{2}
\end{array}\right)
$$

that is an isomorphism from $\mathbb{P}_{2}$ onto $\mathbb{R}^{3}$.

## Coordinate mapping

## Example (continued)

Now we can perform any reasoning in $\mathbb{P}_{2}$ by studying an analogous problem in $\mathbb{R}^{3}$. For instance, let's study if the following polynomials are linearly independent

$$
\left.\begin{array}{lll}
p_{1}(t) & =1+2 t^{2} & \Rightarrow\left[p_{1}(t)\right]_{B}=(1,0,2) \\
p_{2}(t) & =4+t+5 t^{2} & \Rightarrow\left[\begin{array}{l}
2 \\
\left.p_{2}(t)\right]_{B}=(4,1,5) \\
p_{3}(t)
\end{array}=3+2 t\right.
\end{array} \Rightarrow\left[\begin{array}{l}
3
\end{array}\right)(t)\right]_{B}=(3,2,0) .
$$

We simply need to see if the corresponding coordinates in $\mathbb{R}^{3}$ are linearly independent

$$
\left(\begin{array}{lll}
1 & 4 & 3 \\
0 & 1 & 2 \\
2 & 5 & 0
\end{array}\right) \sim\left(\begin{array}{ccc}
1 & 0 & -5 \\
0 & 1 & 2 \\
0 & 0 & 0
\end{array}\right)
$$

Looking at the non-pivot columns we learn that

$$
p_{3}(t)=-5 p_{1}(t)+2 p_{2}(t)
$$

Finally, we conclude that the 3 polynomials are not linearly independent.

## Coordinate mapping

## Example

Consider $\mathbf{v}_{1}=(3,6,2), \mathbf{v}_{2}=(-1,0,1), B=\left\{\mathbf{v}_{1}, \mathbf{v}_{2}\right\}$, and $H=\operatorname{Span}\{B\}$. $H$ is isomorphic to $\mathbb{R}^{2}$ (because its points have only 2 coordinates). For instance, the coordinates of $\mathbf{x}=(3,12,7) \in H$ are $[\mathbf{x}]_{B}=(2,3)$.


FIGURE 7 A coordinate system on a plane $H$ in $\mathbb{R}^{3}$.

## Coordinate mapping

## Example

Consider $\mathbf{v}_{1}=(3,6,2), \mathbf{v}_{2}=(-1,0,1), B=\left\{\mathbf{v}_{1}, \mathbf{v}_{2}\right\}$, and $H=\operatorname{Span}\{B\}$. $H$ is isomorphic to $\mathbb{R}^{2}$ (because its points have only 2 coordinates). For instance, the coordinates of $\mathbf{x}=(3,12,7) \in H$ are $[\mathbf{x}]_{B}=(2,3)$.


FIGURE 7 A coordinate system on a plane $H$ in $\mathbb{R}^{3}$.

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 4, Section 4:

$$
\begin{array}{ll}
\text { - } 4.4 .3 \\
\text { - } & 4.4 .8 \\
\text { - } & 4.4 .9 \\
\text { - } & 4.4 .13 \\
\text { - } 4.4 .17 \\
\text { - } & 4.4 .19 \\
\text { - } & 4.4 .22 \\
\text { - } & 4.4 .24 \\
\text { - } & 4.4 .25
\end{array}
$$

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## Dimension of a vector space

We have just said that if the basis of a vector space $V$ has $n$ elements, then $V$ is isomorphic to $\mathbb{R}^{n}$. $n$ is a characteristic number of each space called the dimension.

## Theorem 9.1

Let $V$ be a vector space with a basis given by $B=\left\{\mathbf{b}_{1}, \mathbf{b}_{2}, \ldots, \mathbf{b}_{n}\right\}$. Then, any subset of $V$ with more than $n$ elements is linearly dependent.
Proof
Let $S$ be a subset of $V$ with $p>n$ vectors

$$
S=\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{p}\right\}
$$

We now consider the set of coordinates of these vectors.

$$
\left\{\left[\mathbf{v}_{1}\right]_{B},\left[\mathbf{v}_{2}\right]_{B}, \ldots,\left[\mathbf{v}_{P}\right]_{B}\right\}
$$

They are $p>n$ vectors in $\mathbb{R}^{n}$ and, therefore, necessarily linearly dependent. That is, there exist $c_{1}, c_{2}, \ldots, c_{p}$, not all of them 0 , such that

$$
c_{1}\left[\mathbf{v}_{1}\right]_{B}+c_{2}\left[\mathbf{v}_{2}\right]_{B}+c_{p}\left[\mathbf{v}_{\rho}\right]_{B}=\mathbf{0} \in \mathbb{R}^{n}
$$

## Dimension of a vector space

Proof (continued)
If we now exploit the fact that the coordinate mapping is linear, then we have

$$
\left[c_{1} \mathbf{v}_{1}+c_{2} \mathbf{v}_{2}+c_{p} \mathbf{v}_{\rho}\right]_{B}=\mathbf{0} \in \mathbb{R}^{n}
$$

Finally, we make use of the fact that the coordinate mapping is bijective

$$
c_{1} \mathbf{v}_{1}+c_{2} \mathbf{v}_{2}+c_{p} \mathbf{v}_{p}=\mathbf{0} \in V
$$

And, consequently, we have shown that the $p$ vectors in $S$ are linearly dependent.

## Theorem 9.2

If a basis of a vector space has $n$ vectors, then all other bases also have $n$ vectors. Proof
Let $B_{1}$ be a basis with $n$ vectors of a vector space $V$. Let $B_{2}$ another basis of $V$. By the previous theorem, $B_{2}$ has at most $n$ vectors. Let us assume now that $B_{2}$ has less than $n$ vectors, then by the previous theorem $B_{1}$ would not be a basis. This is a contradiction with the fact that $B_{1}$ is a basis and, consequently, $B_{2}$ cannot have less than $n$ vectors.

## Dimension of a vector space

## Definition 9.1

If the vector space $V$ is spanned by a finite set of vectors, then $V$ is finite-dimensional and its dimension $(\operatorname{dim}\{V\})$ is the number of elements of any of its bases. The dimension of $V=\{\mathbf{0}\}$ is 0 . If $V$ is not generated by a finite set of vectors, then it is infinite-dimensional.

## Example

```
dim}{\mp@subsup{\mathbb{R}}{}{n}}=
dim}{\mp@subsup{\mathbb{P}}{2}{}}=3\mathrm{ because one of its bases is {1,t,t ' }
dim{\mathbb{P}}=\infty
dim}{\operatorname{Span}{\mp@subsup{\mathbf{v}}{1}{},\mp@subsup{\mathbf{v}}{2}{}}}=
```


## Dimension of a vector space

## Example: in $\mathbb{R}^{3}$

There is a single subspace of dimension $0(\{\mathbf{0}\})$
There are infinite subspaces of dimension 1 (all lines going through the origin)
There are infinite subspaces of dimension 2 (all planes going through the origin)
There is a single subspace of dimension $3\left(\mathbb{R}^{3}\right)$


FIGURE 1 Sample subspaces of $\mathbb{R}^{3}$.

## Dimension of a vector space

## Theorem 9.3

Let $H \subseteq V$ be a vector subspace of a vector space $V$. Then,

$$
\operatorname{dim}\{H\} \leq \operatorname{dim}\{V\}
$$

## Theorem 9.4

Let $V$ a $n$-dimensional vector space ( $n \geq 1$ ).

- Any linearly independent subset of $V$ with $n$ elements is a basis.
- Any subset of $V$ with $n$ elements that span $V$ is a basis.


## Dimension of a vector space

## Theorem 9.5

Consider any matrix $A \in \mathcal{M}_{m \times n}$.

- $\operatorname{dim}\{\operatorname{Nul}\{A\}\}$ is the number of free variables in the equation $A \mathbf{x}=\mathbf{0}$.
- $\operatorname{dim}\{\operatorname{Col}\{A\}\}$ is the number of pivot columns of $A$.


## Example

$$
A=\left(\begin{array}{rrrrr}
-3 & 6 & -1 & 1 & -7 \\
1 & -2 & 2 & 3 & -1 \\
2 & -4 & 5 & 8 & -4
\end{array}\right) \sim\left(\begin{array}{rrrrr}
1 & -2 & 0 & -1 & 3 \\
0 & 0 & 1 & 2 & -2 \\
0 & 0 & 0 & 0 & 0
\end{array}\right)
$$

The number of pivot columns of $A$ is $2=\operatorname{dim}\{\operatorname{Col}\{A\}\}$ (in blue), while the number of free variables is $3=\operatorname{dim}\{\operatorname{Nul}\{A\}\}$ (the free variables are $x_{2}, x_{4}$ and $x_{5}$ ).

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 4, Section 5:

- 4.5.1
- 4.5.13
- 4.5.21
- 4.5.25
- 4.5.26
- 4.5.27
- 4.5.28
- 4.5.31
- 4.5.32


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- Rank of a matrix (d)
- Change of basis (d)


## Rank of a matrix

The rank of a matrix is the number of linearly independent rows of that matrix. It can also be defined as the number of linearly independent columns of that matrix because both definitions yield the same number. We'll see a more formal definition below.

## Definition 10.1 (Row space of a matrix)

Given a matrix $A \in \mathcal{M}_{m \times n}$, the row space of $A$ is the space spanned by all rows of $A\left(\operatorname{Row}\{A\} \subseteq \mathbb{R}^{n}\right)$.

Theorem 10.1

$$
\operatorname{Row}\{A\}=\operatorname{Col}\left\{A^{T}\right\}
$$

## Rank of a matrix

## Theorem 10.2

If a matrix $A$ is row equivalent to another matrix $B$, then $\operatorname{Row}\{A\}=\operatorname{Row}\{B\}$. If $B$ is in a reduced echelon form, then the non-null rows of $B$ form a basis of Row $\{A\}$
Proof

$$
\text { Proof Row }\{A\} \supseteq \operatorname{Row}\{B\}
$$

Since the rows of $B$ are obtained by row operations on the rows of $A$, then any linear combination of the rows of $B$ can be obtained as linear combinations of the rows of $A$.
Proof Row $\{A\} \subseteq \operatorname{Row}\{B\}$
Since the row operations are reversible, then any linear combination of the rows of $A$ can be obtained as linear combinations of the rows of $B$. Proof non-null rows of $B$ form a basis
They are linearly independent because any non-null row of $B$ cannot be obtained as a linear combination of the rows below (because it is in echelon form and there are numbers in early columns that have Os below)

## Rank of a matrix

## Example

$$
A=\left(\begin{array}{rrrrr}
-2 & -5 & 8 & 0 & -17 \\
1 & 3 & -5 & 1 & 5 \\
3 & 11 & -19 & 7 & 1 \\
1 & 7 & -13 & 5 & -3
\end{array}\right) \sim B=\left(\begin{array}{rrrrr}
1 & 3 & -5 & 1 & 5 \\
0 & 1 & -2 & 2 & -7 \\
0 & 0 & 0 & -4 & 20 \\
0 & 0 & 0 & 0 & 0
\end{array}\right)
$$

Pivot columns have been highlighted in blue. At this point we can already construct a basis for the row and column spaces of $A$

$$
\begin{aligned}
\mathbb{R}^{5} \supset \operatorname{Row}\{A\} & =\operatorname{Span}\{(1,3,-5,1,5),(0,1,-2,2,-7),(0,0,0,-4,20)\} \\
\mathbb{R}^{4} \supset \operatorname{Col}\{A\} & =\operatorname{Span}\{(-2,1,3,1),(-5,3,11,7),(0,1,7,5)\}
\end{aligned}
$$

To calculate the null space of $A$ we need the reduced echelon form

$$
A \sim\left(\begin{array}{rrrrr}
1 & 0 & 1 & 0 & 1 \\
0 & 1 & -2 & 0 & 3 \\
0 & 0 & 0 & 1 & -5 \\
0 & 0 & 0 & 0 & 0
\end{array}\right)
$$

## Rank of a matrix

## Example (continued)

$$
\begin{gathered}
A \sim\left(\begin{array}{rrrrr}
1 & 0 & 1 & 0 & 1 \\
0 & 1 & -2 & 0 & 3 \\
0 & 0 & 0 & 1 & -5 \\
0 & 0 & 0 & 0 & 0
\end{array}\right) \Rightarrow \\
x_{1}=-x_{3}-x_{5} \\
x_{2}=2 x_{3}-3 x_{5} \\
x_{4}=5 x_{5}
\end{gathered} \Rightarrow \operatorname{Nul}\{A\} \ni \mathbf{x}=x_{3}\left(\begin{array}{c}
-1 \\
2 \\
1 \\
0 \\
0
\end{array}\right)+x_{5}\left(\begin{array}{c}
-1 \\
-3 \\
0 \\
5 \\
1
\end{array}\right) .
$$

Finally,

$$
\mathbb{R}^{5} \supset \operatorname{Nul}\{A\}=\operatorname{Span}\{(-1,2,1,0,0),(-1,-3,0,5,1)\}
$$

## Rank of a matrix

## Definition 10.2 (Rank of a matrix)

$$
\operatorname{Rank}\{A\}=\operatorname{dim}\{\operatorname{Col}\{A\}\}
$$

That is, by definition, $\operatorname{Rank}\{A\}$ is the number of pivot columns of $A$.

## Rank of a matrix

## Theorem 10.3 (Rank theorem)

For any matrix $A \in \mathcal{M}_{m \times n}$
(1) $\operatorname{dim}\{\operatorname{Row}\{A\}\}=\operatorname{dim}\{\operatorname{Col}\{A\}\}$
(2) $\operatorname{Rank}\{A\}+\operatorname{dim}\{\operatorname{Nul}\{A\}\}=n$

## Proof

(1) Let $B$ be the reduced echelon form of $A$. By definition $\operatorname{Rank}\{A\}$ is the number of pivot columns in $A$ (that is the same as the number of pivot columns in B). Since $B$ is in reduced echelon form, each of its non-zero rows has a column pivot and, consequently, the number of non-zero rows coincides with the number of pivot columns. The basis of $\operatorname{Row}\{B\}=\operatorname{Row}\{A\}$ must have as many elements as pivot columns.
(2) From Theorem 9.5 we know that $\operatorname{Null}\{A\}$ is the number of free variables in $A \mathbf{x}=\mathbf{0}$, that is, the number of non-pivot columns of $B$. Consequently, we have

$$
\operatorname{dim}\{\operatorname{Col}\{A\}\}+\operatorname{dim}\{\operatorname{Nul}\{A\}\}=n
$$

But by definition, $\operatorname{Rank}\{A\}=\operatorname{dim}\{\operatorname{Col}\{A\}\}$, which proves the theorem.

## Rank of a matrix

## Example

Let $A \in \mathcal{M}_{7 \times 9}$. We know $\operatorname{dim}\{\operatorname{Nul}\{A\}\}=2$. What is $\operatorname{Rank}\{A\}$ ? According to the previous theorem

$$
\operatorname{Rank}\{A\}=n-\operatorname{dim}\{\operatorname{Nul}\{A\}\}=9-2=7
$$

## Example

Let $A \in \mathcal{M}_{6 \times 9}$. Can it be $\operatorname{dim}\{\operatorname{Nul}\{A\}\}=2$ ?
Let us presume that it can be $\operatorname{dim}\{\operatorname{Nul}\{A\}\}=2$, then

$$
\operatorname{Rank}\{A\}=n-\operatorname{dim}\{\operatorname{Nul}\{A\}\}=9-2=7
$$

But since $A$ has only 6 rows, the maximum rank can only be 6 (not 7 ), and therefore, it must be $\operatorname{dim}\{\operatorname{Nul}\{A\}\} \geq 3$.

## Rank of a matrix

## Example

$$
A=\left(\begin{array}{rrr}
3 & 0 & -1 \\
3 & 0 & -1 \\
4 & 0 & 5
\end{array}\right) \Rightarrow \begin{aligned}
& \operatorname{Nul}\{A\}=\left\{\left(0, x_{2}, 0\right) \quad \forall x_{2} \in \mathbb{R}\right\} \\
& \operatorname{Row}\{A\}=\left\{\left(x_{1}, 0, x_{3}\right) \quad \forall x_{1}, x_{3} \in \mathbb{R}\right\} \\
& \operatorname{Col}\{A\}=\left\{\left(x_{2}, x_{2}, x_{3}\right) \quad \forall x_{2}, x_{3} \in \mathbb{R}\right\} \\
& \operatorname{Nul}\left\{A^{T}\right\}=\left\{\left(x_{1},-x_{1}, 0\right) \quad \forall x_{1} \in \mathbb{R}\right\}
\end{aligned}
$$



FIGURE 1 Subspaces determined by a matrix $A$.

## Rank of a matrix

## Theorem 10.4 (The invertible matrix theorem (continued))

The following statements are equivalent to those in Theorems 5.1 and 11.5 of Chapter 3 (the invertible matrix theorem). Let $A \in \mathcal{M}_{n \times n}$ xix. The columns of $A$ form a basis of $\mathbb{R}^{n}$.
$x \mathrm{x} . \operatorname{Col}\{A\}=\mathbb{R}^{n}$.
$x x i . \operatorname{dim}\{\operatorname{Col}\{A\}\}=n$
xxii. $\operatorname{Rank}\{A\}=n$
xxiii. $\operatorname{Nul}\{A\}=\{\mathbf{0}\}$.
xxiv. $\operatorname{dim}\{\operatorname{Nul}\{A\}\}=0$.

Proof vii $\Leftrightarrow x x$
vii=The equation $A \mathbf{x}=\mathbf{b}$ has at least one solution for every $\mathbf{b} \in \mathbb{R}^{n}$.
But $\operatorname{Col}\{A\}$ is the set of all $\mathbf{b}$ 's for which $A \mathbf{x}=\mathbf{b}$ has a solution. Therefore, vii $\Rightarrow$ $x x$.
Proof $x x \Leftrightarrow x x i \Leftrightarrow x x i i$
Because of the definition of rank.

## Rank of a matrix

```
Proof \(v\), viii \(\Leftrightarrow x i x\)
\(v \equiv\) The columns of \(A\) are linearly independent.
viii三The columns of \(A\) span \(\mathbb{R}^{n}\).
But both together are the definition of a basis for \(\mathbb{R}^{n}\).
Proof xxi \(\Leftrightarrow\) xxiv
Knowing \(x x i\) and thanks to the rank theorem 10.3 , we can infer that \(\operatorname{dim}\{\operatorname{Nul}\{A\}\}=n-n=0\)
Proof xxiv \(\Leftrightarrow\) xxiii
The only subset with null dimension is \(\{\mathbf{0}\}\).
```


## Exercises

## Exercises

From Lay (3rd ed.), Chapter 4, Section 6:

- 4.6.1
- 4.6.13
- 4.6.15
- 4.6.19
- 4.6.26
- 4.6.28
- 4.6.29
- 4.6.33
- 4.6.35


## Outline

(5) Vector spaces

- Definition (a)
- Vector subspace (a)
- Subspace spanned by a set of vectors (a)
- Null space and column space of a matrix (b)
- Kernel and range of a linear transformation (b)
- Linearly independent sets and bases (b)
- Bases for $\operatorname{Nul}\{A\}$ and $\operatorname{Col}\{A\}$ (c)
- Coordinate system (c)
- Dimension of a vector space (d)
- Rank of a matrix (d)
- Change of basis (d)


## Change of basis

## Example

Let us assume we have a vector $\mathbf{x}$ that has two different coordinates in two different coordinate systems $B$ and $C$.

$$
[\mathbf{x}]_{B}=(3,1) \text { and }[\mathbf{x}]_{C}=(6,4)
$$


(a)

(b)

FIGURE 1 Two coordinate systems for the same vector space.

## Change of basis

## Example (continued)

Presume that for our example

$$
\begin{aligned}
& \mathbf{b}_{1}=4 \mathbf{c}_{1}+\mathbf{c}_{2} \\
& \mathbf{b}_{2}=-6 \mathbf{c}_{1}+\mathbf{c}_{2}
\end{aligned}
$$

We can calculate the coordinates of the basis vectors $B$ in the $C$ coordinate system as

$$
\begin{aligned}
{\left[\mathbf{b}_{1}\right]_{C} } & =(4,1) \\
{\left[\mathbf{b}_{2}\right]_{C} } & =(-6,1)
\end{aligned}
$$

The coordinates of $\mathbf{x}$ in the basis $B$ tell us

$$
\mathbf{x}=3 \mathbf{b}_{1}+\mathbf{b}_{2}
$$

If we now apply the coordinate mapping transformation we have

$$
[\mathbf{x}]_{C}=3\left[\mathbf{b}_{1}\right]_{C}+\left[\mathbf{b}_{2}\right]_{C}=3\binom{4}{1}+\binom{-6}{1}=\left(\begin{array}{rr}
4 & -6 \\
1 & 1
\end{array}\right)\binom{3}{1}=\binom{6}{4}
$$

## Change of basis

## Example (continued)

Note that the columns of the matrix

$$
\left(\begin{array}{rr}
4 & -6 \\
1 & 1
\end{array}\right)
$$

are the coordinates of each one of the elements of the basis $B$ expressed in the coordinate system $C$, and that the overall change of coordinates has the form

$$
[\mathbf{x}]_{C}=\left(\begin{array}{rr}
4 & -6 \\
1 & 1
\end{array}\right)[\mathbf{x}]_{B}
$$

## Change of basis

## Theorem 11.1 (Change of basis)

Let $B=\left\{\mathbf{b}_{1}, \mathbf{b}_{2}, \ldots, \mathbf{b}_{n}\right\}$ and $C=\left\{\mathbf{c}_{1}, \mathbf{c}_{2}, \ldots, \mathbf{c}_{n}\right\}$ be two bases of the vector space $V$. We can transform coordinates from one coordinate system to the other by multiplying by a single, invertible $n \times n$ matrix, called $P_{C \leftarrow B}$ whose columns are the coordinates of the vectors of $B$ in the basis $C$.

$$
[\mathbf{x}]_{C}=P_{C \leftarrow B}[\mathbf{x}]_{B}
$$



## Change of basis

## Corollary

To convert from $C$ coordinates back to $B$ coordinates we simply have to invert the transformation.

$$
P_{B \leftarrow C}=P_{C \leftarrow B}^{-1}
$$

## Corollary

Consider the standard base in $V$ given by $E=\left\{\mathbf{e}_{1}, \mathbf{e}_{2}, \ldots, \mathbf{e}_{n}\right\}$. The matrix to convert the coordinates from $B$ to $E$ is simply

$$
P_{E \leftarrow B}=\left(\begin{array}{llll}
\mathbf{b}_{1} & \mathbf{b}_{2} & \ldots & \mathbf{b}_{n}
\end{array}\right)
$$

Consequently, we have that for two different bases

$$
\mathbf{x}=P_{E \leftarrow B}[\mathbf{x}]_{B}=P_{E \leftarrow C}[\mathbf{x}]_{C}
$$

Finally,

$$
[\mathbf{x}]_{C}=P_{E \leftarrow C}^{-1} P_{E \leftarrow B}[\mathbf{x}]_{B}
$$

## Change of basis

## Numerical trick

Given the two basis $B$ and $C$ we can easily find the coordinates of $B$ in the basis $\mathcal{C}$ in the following way. Let us define two matrices $\mathcal{B}$ and $\mathcal{C}$ whose columns are the elements of the basis. Then

$$
(\mathcal{C} \mid \mathcal{B}) \sim\left(I_{n} \mid P_{\mathcal{C} \leftarrow B}\right)
$$

## Example

Let's say we are given $\mathbf{b}_{1}=(-9,1), \mathbf{b}_{2}=(-5,-1), \mathbf{c}_{1}=(1,-4), \mathbf{c}_{2}=(3,-5)$.

$$
\left(\begin{array}{rr|rr}
1 & 3 & -9 & 5 \\
-4 & -5 & 1 & -1
\end{array}\right) \sim\left(\begin{array}{rr|rr}
1 & 0 & 6 & 4 \\
0 & 1 & -5 & 3
\end{array}\right)
$$

Then, $P_{C \leftarrow B}=\left(\begin{array}{rr}6 & 4 \\ -5 & 3\end{array}\right)$.

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 4, Section 7:

- 4.7.1
- 4.7.9


## Outline

(5) Vector spaces

- Definition (a)
- Vector subspace (a)
- Subspace spanned by a set of vectors (a)
- Null space and column space of a matrix (b)
- Kernel and range of a linear transformation (b)
- Linearly independent sets and bases (b)
- Bases for $\operatorname{Nul}\{A\}$ and $\operatorname{Col}\{A\}$ (c)
- Coordinate system (c)
- Dimension of a vector space (d)
- Rank of a matrix (d)
- Change of basis (d)


# Chapter 6 . Eigenvalues and eigenvectors 

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## Outline

(6) Eigenvalues and eigenvectors

- Definition (a)
- Characteristic equation (a)
- Diagonalization (b)
- Eigenvectors and linear transformations (b)
- Complex eigenvalues (c)


## References


D. Lay. Linear algebra and its applications (3rd ed). Pearson (2006). Chapter 5.

## A little bit of history

Eigenvalues (or "proper values") were first used in the study of the motion of rigid bodies through the inertia matrix by Leonhard Euler and Joseph-Louis Lagrange in the mid of XVIIIth century. Then Augustin-Louis Cauchy used it to analyze quadratic surfaces and conic sections in the early XIXth. Since then, they have found applications in most scientific problems.


## Applications

In this example eigenvalues are used to estimate the size of carotid in a volumetric image.


Hameeteman, K.; Zuluaga, M. A.; et al. Evaluation framework for carotid bifurcation lumen segmentation and stenosis grading. Med Image Anal, 2011, 15, 477-488.

## Applications

In this example eigenvalues were used as a part of another technique (Principal Component Analysis) to automatically analyze luminiscent images.

(a)

(b)

(c)

[^0]
## Outline

(6) Eigenvalues and eigenvectors

- Definition (a)
- Characteristic equation (a)
- Diagonalization (b)
- Eigenvectors and linear transformations (b)
- Complex eigenvalues (c)


## Eigenvalues and eigenvectors

## Example

Consider the linear transformation $T(\mathbf{x})=\left(\begin{array}{cc}3 & -2 \\ 1 & 0\end{array}\right) \mathbf{x}$ on the vectors $\mathbf{u}=(-1,1)$ and $\mathbf{v}=(2,1)$

$$
\begin{aligned}
& T(\mathbf{u})=\left(\begin{array}{cc}
3 & -2 \\
1 & 0
\end{array}\right)\binom{-1}{1}=\binom{-5}{-1} \\
& T(\mathbf{v})=\left(\begin{array}{cc}
3 & -2 \\
1 & 0
\end{array}\right)\binom{2}{1}=\binom{4}{2}
\end{aligned}
$$



FIGURE 1 Effects of multiplication by $A$. $\mathbf{u}$ is changing its direction and module, but $\mathbf{v}$ is only changing its module.

## Eigenvalues and eigenvectors

## Definition 1.1 (Eigenvalue and eigenvector)

Given the matrix $A \in \mathcal{M}_{n \times n}, \lambda$ is an eigenvalue of $A$ if there exists a non-trivial solution $\mathbf{v} \in \mathbb{R}^{n}$ of the equation

$$
A \mathbf{v}=\lambda \mathbf{v}
$$

The solution $\mathbf{v}$ is the eigenvector associated to the eigenvalue $\lambda$.

## Example (continued)

In the previous example, $\mathbf{v}$ was an eigenvector with eigenvalue 2 (because $(2,1) \rightarrow(4,2)$, while $\mathbf{u}$ was not an eigenvector.

## Eigenvalues and eigenvectors

## Example

Show that $\lambda=7$ is an eigenvalue of $A=\left(\begin{array}{ll}1 & 6 \\ 5 & 2\end{array}\right)$.

## Solution

We must find a solution of the equation $A \mathbf{v}=\lambda \mathbf{v}$, or what is the same

$$
\begin{gathered}
A \mathbf{v}-\lambda \mathbf{v}=\mathbf{0} \Rightarrow(A-\lambda /) \mathbf{v}=\mathbf{0} \\
\left(\left(\begin{array}{ll}
1 & 6 \\
5 & 2
\end{array}\right)-7\left(\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right)\right)\binom{v_{1}}{v_{2}}=\left(\begin{array}{cc}
-6 & 6 \\
5 & -5
\end{array}\right)\binom{v_{1}}{v_{2}}=\binom{0}{0}
\end{gathered}
$$

Any vector of the form $\mathbf{v}=\left(v_{1}, v_{1}\right)$ satisfies the previous equation

## Theorem 1.1

In general, eigenvectors are solution of the equation

$$
(A-\lambda /) \mathbf{v}=\mathbf{0}
$$

That is, all eigenvectors belong to $\operatorname{Nul}\{A-\lambda /\}$. This is called the eigenspace.

## Eigenvalues and eigenvectors

## Example (continued)

We see that we have a whole set of vectors associated to $\lambda=7$, this is a subspace of the eigenspace:

$$
\text { Eigenspace }\{7\}=\left\{\left(v_{1}, v_{1}\right) \forall v_{1} \in \mathbb{R}\right\}
$$

It is a line passing through the origin with the direction $(1,1)$.
The other eigenvalue of matrix $A$ is $\lambda=-4$

Eigenspace $\{-4\}=\left\{\left(v_{1},-\frac{5}{6} v_{1}\right) \forall v_{1} \in \mathbb{R}\right\}$


FIGURE 2 Eigenspaces for $\lambda=-4$ and $\lambda=7$.

## Eigenvalues and eigenvectors

## Example

Knowing that $\lambda=2$ is an eigenvalue of $A=\left(\begin{array}{ccc}4 & -1 & 6 \\ 2 & 1 & 6 \\ 2 & -1 & 8\end{array}\right)$, find a basis of its eigenspace.
Solution

$$
A-2 I=\left(\begin{array}{ccc}
4 & -1 & 6 \\
2 & 1 & 6 \\
2 & -1 & 8
\end{array}\right)-\left(\begin{array}{lll}
2 & 0 & 0 \\
0 & 2 & 0 \\
0 & 0 & 2
\end{array}\right)=\left(\begin{array}{lll}
2 & -1 & 6 \\
2 & -1 & 6 \\
2 & -1 & 6
\end{array}\right) \sim\left(\begin{array}{ccc}
2 & -1 & 6 \\
0 & 0 & 0 \\
0 & 0 & 0
\end{array}\right)
$$

So any vector fulfilling this equation must satisfy

$$
x_{1}=\frac{1}{2} x_{2}-3 x_{3} \Rightarrow \text { Eigenspace }\{2\} \ni \mathbf{x}=x_{2}\left(\begin{array}{c}
\frac{1}{2} \\
1 \\
0
\end{array}\right)+x_{3}\left(\begin{array}{c}
-3 \\
0 \\
1
\end{array}\right)
$$

Finally the basis is formed by the vectors $\left(\frac{1}{2}, 1,0\right)$ and $(-3,0,1)$.

## Eigenvalues and eigenvectors

## Example (continued)

Within the eigenspace, $A$ acts as dilation.


## Eigenvalues and eigenvectors

## Theorem 1.2

The eigenvalues of a triangular matrix $A$ are the elements of the main diagonal $\left(a_{i i}, i=1,2, \ldots, n\right)$.
Proof
Consider the matrix $A-\lambda /$

$$
\left(\begin{array}{ccccc}
a_{11}-\lambda & a_{12} & a_{13} & \ldots & a_{1 n} \\
0 & a_{22}-\lambda & a_{23} & \ldots & a_{2 n} \\
0 & 0 & a_{33}-\lambda & \ldots & a_{3 n} \\
\ldots & \ldots & \ldots & \ldots & \ldots \\
0 & 0 & 0 & \ldots & a_{n n}-\lambda
\end{array}\right)
$$

The equation system $A-\lambda I=\mathbf{0}$ has a non-trivial solution if at least 1 of the entries in the diagonal is 0 . Therefore, it must be $\lambda=a_{i i}$ for some $i$. Varying $i$ from 1 to $n$ we obtain that all the elements in the main diagonal are the $n$ eigenvalues of the matrix $A$.

## Eigenvalues and eigenvectors

## Example

The eigenvalues of $A=\left(\begin{array}{ccc}3 & 6 & -8 \\ 0 & 0 & 6 \\ 0 & 0 & 2\end{array}\right)$ are $\lambda=3,0,2$.

## Theorem 1.3

Let $\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{r}$ be $r$ eigenvectors associated to $r$ different eigenvalues. Then, the set $S=\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{r}\right\}$ is linearly independent.
Proof
Let us assume that $S$ is linearly dependent. Without loss of generality, we may assume that the first $p(p<r)$ are linearly independent, and that the $p+1$-th vector is dependent on the precedent vectors. Then, there must exist $c_{1}, c_{2}, \ldots, c_{p}$ not all of them zero such that

$$
\begin{equation*}
\mathbf{v}_{p+1}=c_{1} \mathbf{v}_{1}+c_{2} \mathbf{v}_{2}+\ldots+c_{p} \mathbf{v}_{p} \tag{1}
\end{equation*}
$$

## Eigenvalues and eigenvectors

If we multiply both sides of the equation by $A$, then we have

$$
\begin{align*}
A \mathbf{v}_{p+1} & =c_{1} A \mathbf{v}_{1}+c_{2} A \mathbf{v}_{2}+\ldots+c_{p} A \mathbf{v}_{p}  \tag{2}\\
\lambda_{p+1} \mathbf{v}_{p+1} & =c_{1} \lambda_{1} \mathbf{v}_{1}+c_{2} \lambda_{2} \mathbf{v}_{2}+\ldots+c_{p} \lambda_{p} \mathbf{v}_{p}
\end{align*}
$$

If we multiply Eq. (1) by $\lambda_{p+1}$ and subtract from Eq. (2), we have

$$
\mathbf{0}=c_{1}\left(\lambda_{1}-\lambda_{p+1}\right) \mathbf{v}_{1}+c_{2}\left(\lambda_{2}-\lambda_{p+1}\right) \mathbf{v}_{2}+\ldots+c_{p}\left(\lambda_{p}-\lambda_{p+1}\right) \mathbf{v}_{p}
$$

Since the first $p$ vectors are linearly independent it must be for $i=1,2, \ldots, p$

$$
c_{i}\left(\lambda_{i}-\lambda_{p+1}\right)=0
$$

Because all eigenvalues are different, then it must be $c_{i}=0(i=1,2, \ldots, p)$. But this is a contradiction with the initial hypothesis that not all of them were 0 . Consequently, the set $S$ must be linearly independent. (q.e.d.)

## Eigenvalues and eigenvectors

## Difference equations

Let us assume we have two populations of cells: stem cells and mature cells. Everyday we measure the number of them and we observe that:

## Stem cells:

- $80 \%$ of them have remained as stem cells
- $15 \%$ of them have differentiated into somatic cells


## Somatic cells:

- $95 \%$ of them have remained as somatic cells
- $5 \%$ of them have died
- $5 \%$ of them have died
- There are $20 \%$ new stem cells.



## Eigenvalues and eigenvectors

## Difference equations (continued)

If we call $x_{\text {stem }}^{(k)}$ the number of stem cells on the day $k$, and $x_{\text {somatic }}^{(k)}$ the number of somatic cells the same day, then the following equation reflects the dynamics of the system:

$$
\binom{x_{\text {stem }}^{(k+1)}}{x_{\text {somatic }}^{(k+1)}}=\left(\begin{array}{cc}
1 & 0 \\
0.15 & 0.95
\end{array}\right)\binom{x_{\text {stem }}^{(k)}}{x_{\text {somatic }}^{(k)}}
$$

Let us assume that the day 0 , there are 10,000 stem cells, and 0 somatic cells. Then, the evolution over time is

$$
\begin{aligned}
& \binom{x_{\text {stem }}^{(1)}}{x_{\text {somatic }}^{(1)}}=\left(\begin{array}{cc}
1 & 0 \\
0.15 & 0.95
\end{array}\right)\binom{x_{\text {stem }}^{(0)}}{x_{\text {somatic }}^{(0)}}=\left(\begin{array}{cc}
1 & 0 \\
0.15 & 0.95
\end{array}\right)\binom{10,000}{0}=\binom{10,000}{1,500} \\
& \binom{x_{\text {stem }}^{(2)}}{x_{\text {somatic }}^{(2)}}=\left(\begin{array}{cc}
1 & 0 \\
0.15 & 0.95
\end{array}\right)\binom{x_{\text {stem }}^{(1)}}{x_{\text {somatic }}^{(1)}}=\left(\begin{array}{cc}
1 & 0 \\
0.15 & 0.95
\end{array}\right)\binom{10,000}{1,500}=\binom{10,000}{2,925}
\end{aligned}
$$

## Eigenvalues and eigenvectors

## Difference equations

The previous model is of the form

$$
\mathbf{x}^{(k+1)}=A \mathbf{x}^{(k)}
$$

The simplest way of constructing a solution of the previous equation is by taking an eigenvector $\mathbf{x}_{1}$ and its corresponding eigenvalue, $\lambda$ :

$$
\mathbf{x}^{(k)}=\lambda_{1}^{k} \mathbf{x}_{1}
$$

This is actually a solution because:

$$
\mathbf{x}^{(k+1)}=A \mathbf{x}^{(k)}=A\left(\lambda_{1}^{k} \mathbf{x}_{1}\right)=\lambda_{1}^{k}\left(A \mathbf{x}_{1}\right)=\lambda_{1}^{k}\left(\lambda_{1} \mathbf{x}_{1}\right)=\lambda_{1}^{k+1} \mathbf{x}_{1}
$$

It turns out that any linear combination of eigenvectors is also a solution

$$
\mathbf{x}^{(k)}=c_{1} \lambda_{1}^{k} \mathbf{x}_{1}+c_{2} \lambda_{2}^{k} \mathbf{x}_{2}+\ldots+c_{n} \lambda_{n}^{k} \mathbf{x}_{n}
$$

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 5, Section 1:

- 5.1.1
- 5.1.3
- 5.1.9
- 5.1.17
- 5.1.19
- 5.1.23
- 5.1.25
- 5.1.26
- 5.1.27


## Outline

(6) Eigenvalues and eigenvectors

- Definition (a)
- Characteristic equation (a)
- Diagonalization (b)
- Eigenvectors and linear transformations (b)
- Complex eigenvalues (c)


## Characteristic equation

## Example

Find the eigenvalues of $A=\left(\begin{array}{cc}2 & 3 \\ 3 & -6\end{array}\right)$

## Solution

We need to find scalar values $\lambda$ such that the equation

$$
(A-\lambda /) \mathbf{x}=\mathbf{0}
$$

has non-trivial solutions. By the Invertible Matrix theorem we know that this problem is equivalent to that of finding $\lambda$ values such that

$$
|A-\lambda I|=0
$$

In this case

$$
\begin{aligned}
& \left|\left(\begin{array}{cc}
2 & 3 \\
3 & -6
\end{array}\right)-\left(\begin{array}{cc}
\lambda & 0 \\
0 & \lambda
\end{array}\right)\right|=0 \\
& \mid=(2-\lambda)(-6-\lambda)-9=\lambda^{2}+4 \lambda-21=0
\end{aligned}
$$

## Characteristic equation

## Example (continued)

$$
\lambda^{2}+4 \lambda-21=0 \Rightarrow \lambda=\frac{-4 \pm \sqrt{4^{2}-4 \cdot 1 \cdot(-21)}}{2 \cdot 1}=\left\{\begin{array}{c}
-7 \\
3
\end{array}\right.
$$

## Theorem 2.1 (The invertible matrix theorem (continued))

This theorem adds to the Theorems 5.1, 11.5 of Chapter 3 and 10.4 of Chapter 5.
$x \times v .|A| \neq 0$.
xxvi. 0 is not an eigenvalue of $A$.

## Definition 2.1 (Characteristic equation)

$A$ scalar $\lambda$ is an eigenvalue of a matrix $A \in \mathcal{M}_{n \times n}$ iff it is solution of the characteristic equation

$$
|A-\lambda I|=0
$$

The determinant of $A-\lambda /$ is called the characteristic polynomial.

## Characteristic equation

## Example

Let us calculate the eigenvalues of $A=\left(\begin{array}{cccc}5 & -2 & 6 & -1 \\ 0 & 3 & -8 & 0 \\ 0 & 0 & 5 & 4 \\ 0 & 0 & 0 & 1\end{array}\right)$.

$$
|A-\lambda I|=\left|\begin{array}{cccc}
5-\lambda & -2 & 6 & -1 \\
0 & 3-\lambda & -8 & 0 \\
0 & 0 & 5-\lambda & 4 \\
0 & 0 & 0 & 1-\lambda
\end{array}\right|=(5-\lambda)^{2},(3-\lambda)(1-\lambda)=0
$$

whose solutions are $\lambda=5$ (with multiplicity 2 ), $\lambda=3$, and $\lambda=1$.

## Example

Let us find the eigenvalues of a matrix whose characteristic polynomial is

$$
|A-\lambda I|=\lambda^{6}-4 \lambda^{5}-12 \lambda^{4}=\lambda^{4}\left(\lambda^{2}-4 \lambda-12\right)=\lambda^{4}(\lambda-6)(\lambda+2)=0
$$

whose solutions are $\lambda=0$ (with multiplicity 4 ), $\lambda=6$, and $\lambda=-2$.

## Characteristic equation

## Definition 2.2 (Similarity between matrices)

Given two matrices $A, B \in \mathcal{M}_{n \times n}$, $A$ is similar to $B$ iff there exists an invertible matrix $P \in \mathcal{M}_{n \times n}$ such that

$$
B=P^{-1} A P
$$

Watch out that similarity is not the same as row equivalence ( $A$ and $B$ are row equivalent if there exists a $E$ such that $B=E A$ being $E$ invertible and the product of row operation matrices).

## Characteristic equation

## Theorem 2.2

If $A$ is similar to $B$, then $B$ is similar to $A$.
Proof
It suffices to take the definition of $A$ similar to $B$ and solve for $B$. If we multiply by $P$ on the right

$$
B=P^{-1} A P \Rightarrow P B=A P
$$

Now, we multiply by $P$ on the left ( $P^{-1}$ exists because $P$ is invertible)

$$
P B=A P \Rightarrow P B P^{-1}=A
$$

and this is the definition of $B$ being similar to $A$.

## Characteristic equation

## Theorem 2.3

If $A$ and $B$ are similar matrices, then they have the same characteristic polynomial. Proof
If $A$ is similar to $B$, then there exists an invertible matrix $P$ such that

$$
B=P^{-1} A P
$$

If we subtract on both sides $\lambda /$ we have

$$
B-\lambda I=P^{-1} A P-\lambda I=P^{-1} A P-\lambda P^{-1} P=P^{-1}(A-\lambda I) P
$$

Now taking the determinant of both sides

$$
|B-\lambda I|=\left|P^{-1}(A-\lambda I) P\right|=\left|P^{-1}\right||A-\lambda I||P|=|P|^{-1}|A-\lambda I||P|=|A-\lambda I|
$$

## Characteristic equation

## Theorem 2.4

If $A$ and $B$ are similar matrices, then they have the same characteristic polynomial. Proof
If $A$ is similar to $B$, then there exists an invertible matrix $P$ such that

$$
B=P^{-1} A P
$$

If we subtract on both sides $\lambda /$ we have

$$
B-\lambda I=P^{-1} A P-\lambda I=P^{-1} A P-\lambda P^{-1} P=P^{-1}(A-\lambda I) P
$$

Now taking the determinant of both sides

$$
|B-\lambda I|=\left|P^{-1}(A-\lambda I) P\right|=\left|P^{-1}\right||A-\lambda I||P|=|P|^{-1}|A-\lambda I||P|=|A-\lambda I|
$$

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 5, Section 2:

- 5.2.1
- 5.2.9
- 5.2.18
- 5.2.19
- 5.2.20
- 5.2.23
- 5.2.24
- 5.2.28 (computer)


## Outline

(6) Eigenvalues and eigenvectors

- Definition (a)
- Characteristic equation (a)
- Diagonalization (b)
- Eigenvectors and linear transformations (b)
- Complex eigenvalues (c)


## Diagonalization

## Definition 3.1 (Diagonalization)

$A \in \mathcal{M}_{n \times n}$ is diagonalizable if there exists $P, D \in \mathcal{M}_{n \times n}$ (with $P$ invertible and
$D$ diagonal) such that

$$
A=P D P^{-1}
$$

Diagonalization simplifies the calculation of powers of $A\left(A^{k}\right)$, is used to decouple dynamic systems, and in multivariate statistics to produce uncorrelated random variables.

## Example

$$
D=\left(\begin{array}{ll}
5 & 0 \\
0 & 3
\end{array}\right) \quad D^{2}=\left(\begin{array}{cc}
5^{2} & 0 \\
0 & 3^{2}
\end{array}\right) \quad D^{3}=\left(\begin{array}{cc}
5^{3} & 0 \\
0 & 3^{3}
\end{array}\right)
$$

## Diagonalization

## Example

Let us assume that $A=P D P^{-1}$. Let us calculate calculate now the different powers of $A$

$$
\begin{aligned}
& A^{2}=A \cdot A=\left(P D P^{-1}\right)\left(P D P^{-1}\right)=(P D)\left(P^{-1} P\right)\left(D P^{-1}\right)=P D D P^{-1}=P D^{2} P^{-1} \\
& A^{3}=A^{2} \cdot A=\left(P D^{2} P^{-1}\right)\left(P D P^{-1}\right)=P D^{3} P^{-1} \\
& \cdots \\
& A^{k}=P D^{k} P^{-1}
\end{aligned}
$$

Let us particularize this result for $A=\left(\begin{array}{cc}7 & 2 \\ -4 & 1\end{array}\right)$ that can be factorized with $P=\left(\begin{array}{cc}1 & 1 \\ -1 & -2\end{array}\right)$ and $D=\left(\begin{array}{ll}5 & 0 \\ 0 & 3\end{array}\right)$ as $A=P D P^{-1}$.

$$
\begin{gathered}
A^{k}=P D^{k} P^{-1}=\left(\begin{array}{cc}
1 & 1 \\
-1 & -2
\end{array}\right)\left(\begin{array}{cc}
5^{k} & 0 \\
0 & 3^{k}
\end{array}\right)\left(\begin{array}{cc}
2 & 1 \\
-1 & -1
\end{array}\right)= \\
\left(\begin{array}{cc}
2 \cdot 5^{k}-3^{k} & 5^{k}-3^{k} \\
2 \cdot 3^{k}-2 \cdot 5^{k} & 2 \cdot 3^{k}-5^{k}
\end{array}\right)
\end{gathered}
$$

## Diagonalization

## Theorem 3.1 (Diagonalization theorem)

$A \in \mathcal{M}_{n \times n}$ is diagonalizable iff $A$ has $n$ linearly independent eigenvectors. In this case, we may construct $P$ by stacking the $n$ eigenvectors, and $D$ as a diagonal matrix with the corresponding eigenvalues.
Proof
Consider the columns of $P=\left(\begin{array}{llll}\mathbf{p}_{1} & \mathbf{p}_{2} & \ldots & \mathbf{p}_{n}\end{array}\right)$ and $D=\left(\begin{array}{cccc}d_{1} & 0 & \ldots & 0 \\ 0 & d_{2} & \ldots & 0 \\ \ldots & \ldots & \ldots & \ldots \\ 0 & 0 & \ldots & d_{n}\end{array}\right)$
Let us assume that $A=P D P^{-1}$ and we multiply by $P$ on the right

$$
\left.\begin{array}{rl}
A P & =P D \\
A\left(\begin{array}{llll}
\mathbf{p}_{1} & \mathbf{p}_{2} & \ldots & \mathbf{p}_{n}
\end{array}\right) & =\left(\begin{array}{lllll}
\mathbf{p}_{1} & \mathbf{p}_{2} & \ldots & \mathbf{p}_{n}
\end{array}\right)\left(\begin{array}{cccc}
d_{1} & 0 & \ldots & 0 \\
0 & d_{2} & \ldots & 0 \\
\ldots & \ldots & \ldots & \ldots \\
0 & 0 & \ldots & d_{n}
\end{array}\right) \\
A \mathbf{p}_{1} & A \mathbf{p}_{2}
\end{array} \ldots \quad A \mathbf{p}_{n}\right) ~=\left(\begin{array}{lllll}
d_{1} \mathbf{p}_{1} & d_{2} \mathbf{p}_{2} & \ldots & d_{n} \mathbf{p}_{n}
\end{array}\right)
$$

## Diagonalization

This implies that

$$
\begin{gathered}
A \mathbf{p}_{1}=d_{1} \mathbf{p}_{1} \\
A \mathbf{p}_{2}=d_{2} \mathbf{p}_{2} \\
\ldots \\
A \\
A \mathbf{p}_{n}=d_{n} \mathbf{p}_{n}
\end{gathered}
$$

But this is the definition of eigenvector, so the columns of $P\left(\mathbf{p}_{i}\right)$ must be eigenvectors of $A$ and $d_{i}$ its corresponding eigenvalue. Since $P$ is invertible, its columns must be linearly independent.

## Diagonalization

## Example

Diagonalize $A=\left(\begin{array}{ccc}1 & 3 & 3 \\ -3 & -5 & -3 \\ 3 & 3 & 1\end{array}\right)$.
Step 1: Find the eigenvalues of $A$

$$
|A-\lambda I|=0 \Rightarrow-\lambda^{3}-3 \lambda^{2}+4=-(\lambda-1)(\lambda+2)^{2}=0
$$

whose solutions are $\lambda=1$ and $\lambda=-2$ (double).
Step 2: Find a linearly independent set of eigenvectors
$\underline{\lambda=1}$
$A-\lambda I=\left(\begin{array}{ccc}1 & 3 & 3 \\ -3 & -5 & -3 \\ 3 & 3 & 1\end{array}\right)-\left(\begin{array}{ccc}1 & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 0 & 1\end{array}\right)=\left(\begin{array}{ccc}0 & 3 & 3 \\ -3 & -6 & -3 \\ 3 & 3 & 0\end{array}\right) \sim\left(\begin{array}{lll}0 & 1 & 1 \\ 0 & 0 & 0 \\ 1 & 1 & 0\end{array}\right)$

## Diagonalization

## Example (continued)

Step 2: Find a linearly independent set of eigenvectors
$\lambda=1$

$$
A-\lambda I \sim\left(\begin{array}{lll}
0 & 1 & 1 \\
0 & 0 & 0 \\
1 & 1 & 0
\end{array}\right) \Rightarrow \begin{aligned}
& x_{1}=-x_{2} \\
& x_{3}=-x_{2}
\end{aligned} \Rightarrow \mathbf{v}_{1}=\left(\begin{array}{c}
1 \\
-1 \\
1
\end{array}\right)
$$

$\lambda=-2$

$$
\begin{gathered}
A-\lambda I=\left(\begin{array}{ccc}
1 & 3 & 3 \\
-3 & -5 & -3 \\
3 & 3 & 1
\end{array}\right)-\left(\begin{array}{ccc}
-2 & 0 & 0 \\
0 & -2 & 0 \\
0 & 0 & -2
\end{array}\right)=\left(\begin{array}{ccc}
3 & 3 & 3 \\
-3 & -3 & -3 \\
3 & 3 & 3
\end{array}\right) \sim \\
\left(\begin{array}{lll}
1 & 1 & 1 \\
0 & 0 & 0 \\
0 & 0 & 0
\end{array}\right) \Rightarrow x_{1}=-x_{2}-x_{3} \Rightarrow \mathbf{v}_{2}=\left(\begin{array}{c}
-1 \\
1 \\
0
\end{array}\right), \mathbf{v}_{3}=\left(\begin{array}{c}
-1 \\
0 \\
1
\end{array}\right)
\end{gathered}
$$

## Diagonalization

## Example (continued)

Step 3: Construct $P$ and $D$

$$
P=\left(\begin{array}{ccc}
1 & -1 & -1 \\
-1 & 1 & 0 \\
1 & 0 & 1
\end{array}\right) \quad D=\left(\begin{array}{ccc}
1 & 0 & 0 \\
0 & -2 & 0 \\
0 & 0 & -2
\end{array}\right)
$$

Step 4: Check everything is correct
$P$ is invertible $|P| \neq 0$

$$
|P|=1
$$

$$
A=P D P^{-1} \Rightarrow A P=P D
$$

$$
A P=\left(\begin{array}{ccc}
1 & 2 & 2 \\
-1 & -2 & 0 \\
1 & 0 & -2
\end{array}\right) \quad P D=\left(\begin{array}{ccc}
1 & 2 & 2 \\
-1 & -2 & 0 \\
1 & 0 & -2
\end{array}\right)
$$

## Diagonalization

## Example (continued)

Step 4: Check everything is correct
$P$ is invertible $|P| \neq 0$
MATLAB:

$$
\begin{aligned}
& \mathrm{P}=\left[\begin{array}{lllllllll}
1 & -1 & -1 ; & -1 & 1 & 0 & 1 & 0 & 1
\end{array}\right] ; \\
& \operatorname{det}(\mathrm{P}) \\
& \quad A=P D P^{-1} \Rightarrow A P=P D \\
& \mathrm{MATLAB}
\end{aligned} \mathrm{~A}=\left[\begin{array}{lllllllll}
1 & 3 & 3 ; & -3 & -5 & 3 ; & 3 & 3 & 1
\end{array}\right] ; \text {; }
$$

## Diagonalization

## Example

Diagonalize $A=\left(\begin{array}{ccc}2 & 4 & 3 \\ -4 & -6 & -3 \\ 3 & 3 & 1\end{array}\right)$.
Step 1: Find the eigenvalues of $A$

$$
|A-\lambda I|=0 \Rightarrow-\lambda^{3}-3 \lambda^{2}+4=-(\lambda-1)(\lambda+2)^{2}=0
$$

whose solutions are $\lambda=1$ and $\lambda=-2$ (double). (Same eigenvalues as in the previous example)
Step 2: Find a linearly independent set of eigenvectors
$\underline{\lambda=1}$
$A-\lambda I=\left(\begin{array}{ccc}2 & 4 & 3 \\ -4 & -6 & -3 \\ 3 & 3 & 1\end{array}\right)-\left(\begin{array}{ccc}1 & 0 & 0 \\ 0 & 1 & 0 \\ 0 & 0 & 1\end{array}\right)=\left(\begin{array}{ccc}1 & 4 & 3 \\ -4 & -7 & -3 \\ 3 & 3 & 0\end{array}\right) \sim\left(\begin{array}{ccc}1 & 0 & -1 \\ 0 & 1 & 1 \\ 0 & 0 & 0\end{array}\right)$

## Diagonalization

## Example (continued)

Step 2: Find a linearly independent set of eigenvectors
$\underline{\lambda=1}$

$$
A-\lambda I \sim\left(\begin{array}{ccc}
1 & 0 & -1 \\
0 & 1 & 1 \\
0 & 0 & 0
\end{array}\right) \Rightarrow \begin{gathered}
x_{1}=x_{3} \\
x_{2}=-x_{3}
\end{gathered} \Rightarrow \mathbf{v}_{1}=\left(\begin{array}{c}
1 \\
-1 \\
1
\end{array}\right)
$$

(The same eigenspace as in the previous example).

$$
\lambda=-2
$$

$$
\begin{aligned}
A-\lambda I= & \left(\begin{array}{ccc}
2 & 4 & 3 \\
-4 & -6 & -3 \\
3 & 3 & 1
\end{array}\right)-\left(\begin{array}{ccc}
-2 & 0 & 0 \\
0 & -2 & 0 \\
0 & 0 & -2
\end{array}\right)=\left(\begin{array}{ccc}
4 & 4 & 3 \\
-4 & -4 & -3 \\
3 & 3 & 3
\end{array}\right) \sim \\
& \left(\begin{array}{ccc}
1 & 1 & \frac{3}{4} \\
0 & 0 & 0 \\
0 & 0 & \frac{1}{4}
\end{array}\right) \Rightarrow \begin{array}{c}
x_{1}=-x_{2}-\frac{3}{4} x_{3} \\
\frac{1}{4} x_{3}=0
\end{array} \Rightarrow \mathbf{v}_{2}=\left(\begin{array}{c}
-1 \\
1 \\
0
\end{array}\right)
\end{aligned}
$$

( $A$ cannot be diagonalized because there are not 3 linearly independent vectors)

## Diagonalization

## Theorem 3.2

If a $n \times n$ matrix has $n$ different eigenvalues, then it is diagonalizable. Proof
Let $\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{n}$ be the $n$ eigenvectors corresponding to the $n$ different eigenvalues. The set

$$
\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{n}\right\}
$$

is linearly independent by Theorem 1.3 and $A$ is diagonalizable by Theorem 3.1.

## Example

Is $A=\left(\begin{array}{ccc}5 & -8 & 1 \\ 0 & 0 & 7 \\ 0 & 0 & -2\end{array}\right)$ diagonalizable?

## Solution

$A$ is a triangular matrix and its eigenvalues are 5,0 and -2 , all of them distinct, and by the previous theorem $A$ is diagonalizable.

## Diagonalization

## Theorem 3.3

Let $A \in \mathcal{M}_{n \times n}$ with $p \leq n$ different eigenvalues. Let $d_{k}$ be the dimension associated to the eigenvalue $\lambda_{k}$. Then,
(1) $d_{k}$ is smaller or equal the multiplicity of $\lambda_{k}$.
(2) $A$ is diagonalizable iff $d_{k}$ is equal to the multiplicity of $\lambda_{k}$. In this case,

$$
\sum_{k=1}^{p} d_{k}=n
$$

(3) If $A$ is diagonalizable and $B_{k}$ are the bases of each one of the eigenspaces, then $\left\{B_{1}, B_{2}, \ldots, B_{p}\right\}$ is a basis of $\mathbb{R}^{n}$.

## Diagonalization

## Example

Let $A=\left(\begin{array}{cccc}5 & 0 & 0 & 0 \\ 0 & 5 & 0 & 0 \\ 1 & 4 & -3 & 0 \\ -1 & -2 & 0 & 3\end{array}\right)$. Let's factorize it as $A=P D P^{-1}$. The eigenvalues and associated eigenvectors are

$$
\begin{aligned}
\lambda_{1}=5 & \leftrightarrow \mathbf{v}_{1}=\left(\begin{array}{c}
-8 \\
4 \\
1 \\
0
\end{array}\right)
\end{aligned} \mathbf{v}_{2}=\left(\begin{array}{c}
-16 \\
4 \\
0 \\
1
\end{array}\right) \quad \Rightarrow \quad \begin{aligned}
& 0=\left(\begin{array}{cccc}
-8 & -16 & 0 & 0 \\
4 & 4 & 0 & 0 \\
1 & 0 & 1 & 0 \\
0 & 1 & 0 & 1
\end{array}\right) \\
& \lambda_{2}=-3
\end{aligned} \quad \mathbf{v}_{3}=\left(\begin{array}{l}
0 \\
0 \\
1 \\
0
\end{array}\right) \quad \mathbf{v}_{4}=\left(\begin{array}{cccc}
5 & 0 & 0 & 0 \\
0 \\
0 & 5 & 0 & 0 \\
0 & 0 & -3 & 0 \\
0 & 0 & 0 & -3
\end{array}\right) .
$$

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 5, Section 3:

- 5.3.1
- 5.3.23
- 5.3.27
- 5.3.28
- 5.3.29
- 5.3.31
- 5.3.32
- 5.3.33 (computer)


## Outline

(6) Eigenvalues and eigenvectors

- Definition (a)
- Characteristic equation (a)
- Diagonalization (b)
- Eigenvectors and linear transformations (b)
- Complex eigenvalues (c)


## The matrix of a linear transformation

The objective of this section is to show that if $A$ is diagonalizable $\left(A=P D P^{-1}\right)$, then the transformation $T_{A}(\mathbf{x})=A \mathbf{x}$ is essentially the same as $T_{D}(\mathbf{u})=D \mathbf{u}$.

## Definition 4.1 (The matrix of a linear transformation)

Consider a linear transformation between two vectors spaces $T: U \rightarrow V$. Let $B$ be a basis of $V$, and $C$ be a basis of $W$. Let $\mathbf{x} \in V$ and consider its coordinates $[\mathrm{x}]_{B}=\left(r_{1}, r_{2}, \ldots, r_{n}\right)$.


FIGURE 1 A linear transformation from $V$ to $W$.

## The matrix of a linear transformations

Let's analyze $\mathbf{x}$ and $T(\mathbf{x})$

$$
\begin{aligned}
\mathbf{x} & =r_{1} \mathbf{b}_{1}+r_{2} \mathbf{b}_{2}+\ldots+r_{n} \mathbf{b}_{n} \Rightarrow \\
T(\mathbf{x}) & =T\left(r_{1} \mathbf{b}_{1}+r_{2} \mathbf{b}_{2}+\ldots+r_{n} \mathbf{b}_{n}\right) \text { [T is linear] } \\
& =r_{1} T\left(\mathbf{b}_{1}\right)+r_{2} T\left(\mathbf{b}_{2}\right)+\ldots+r_{n} T\left(\mathbf{b}_{n}\right)
\end{aligned}
$$

Now, let us consider the coordinates in $C$ of the transformed vector

$$
[T(\mathbf{x})]_{C}=r_{1}\left[T\left(\mathbf{b}_{1}\right)\right]_{C}+r_{2}\left[T\left(\mathbf{b}_{2}\right)\right]_{C}+\ldots+r_{n}\left[T\left(\mathbf{b}_{n}\right)\right]_{C}
$$

We can write this equation in matrix form as

$$
[T(\mathbf{x})]_{C}=M[\mathbf{x}]_{B}
$$

where $M \in \mathcal{M}_{m \times n}$ is a matrix formed by the transformations of each one of the basis vectors in $B$

$$
M=\left([ \begin{array} { l l l l } 
{ T ( \mathbf { b } _ { 1 } ) }
\end{array} ] c \quad \left[\begin{array}{lll}
\left.T\left(\mathbf{b}_{2}\right)\right]_{c} & \ldots & \left.\left[T\left(\mathbf{b}_{n}\right)\right]_{c}\right)
\end{array}\right.\right.
$$

Matrix $M$ is called the matrix of $T$ relative to the bases $B$ and $C$.

## The matrix of a linear transformations



## Example

Let $B=\left\{\mathbf{b}_{1}, \mathbf{b}_{2}\right\}$ and $C=\left\{\mathbf{c}_{1}, \mathbf{c}_{2}, \mathbf{c}_{3}\right\}$ and

$$
\begin{aligned}
& T\left(\mathbf{b}_{1}\right)=3 \mathbf{c}_{1}-2 \mathbf{c}_{2}+5 \mathbf{c}_{3} \\
& T\left(\mathbf{b}_{2}\right)=4 \mathbf{c}_{1}+7 \mathbf{c}_{2}-\mathbf{c}_{3}
\end{aligned} \Rightarrow M=\left(\begin{array}{cc}
3 & 4 \\
-2 & 7 \\
5 & -1
\end{array}\right)
$$

## Transformations from $V$ into $V$

## Definition 4.2 ( $B$-matrix for $T$ )

If $T$ is a transformation from $V$ into $V$ and $B$ is a basis of $V$, then the matrix $M$ is called the $B$-matrix of $T$.

## Example

Consider in the vector space of polynomials of degree $2\left(\mathbb{P}_{2}\right)$, the derivative transformation

$$
\begin{aligned}
T: & \mathbb{P}_{2} \rightarrow \mathbb{P}_{2} \\
& T\left(a_{0}+a_{1} t+a_{2} t^{2}\right)=a_{1}+2 a_{2} t
\end{aligned}
$$

Consider the standard basis of $\mathbb{P}_{2}, B=\left\{1, t, t^{2}\right\}$.

## Transformations from $V$ into $V$

## Example (continued)

Which is the $B$-transformation matrix?
Solution

$$
\begin{aligned}
& T(1)=0 \rightarrow[T(1)]_{B}=\left(\begin{array}{l}
0 \\
0 \\
0
\end{array}\right) \\
& T(t)=1 \rightarrow[T(t)]_{B}=\left(\begin{array}{l}
1 \\
0 \\
0
\end{array}\right) \Rightarrow M=\left(\begin{array}{lll}
0 & 1 & 0 \\
0 & 0 & 2 \\
0 & 0 & 0
\end{array}\right) \\
& T\left(t^{2}\right)=2 t \rightarrow\left[T\left(t^{2}\right)\right]_{B}=\left(\begin{array}{l}
0 \\
2 \\
0
\end{array}\right)
\end{aligned}
$$

## Transformations from $V$ into $V$

## Example (continued)

Verify that $[T(\mathbf{x})]_{B}=M[\mathbf{x}]_{B}$

## Solution

Given any polynomial $p(t)=a_{0}+a_{1} t+a_{2} t^{2}$ its coordinates are $[p(t)]_{B}=\left(a_{0}, a_{1}, a_{2}\right)$. The derivative of $p(t)$ is $T(p(t))=a_{1}+2 a_{2} t$, then

$$
[T(p(t))]_{B}=\left(\begin{array}{c}
a_{1} \\
2 a_{2} \\
0
\end{array}\right)=\left(\begin{array}{lll}
0 & 1 & 0 \\
0 & 0 & 2 \\
0 & 0 & 0
\end{array}\right)\left(\begin{array}{c}
a_{0} \\
a_{1} \\
a_{2}
\end{array}\right)
$$



## Transformations from $\mathbb{R}^{n}$ into $\mathbb{R}^{n}$

## Theorem 4.1 (Diagonal matrix representation)

Suppose matrix $A$ is diagonalizable $\left(A=P D P^{-1}\right)$. If $B$ is the basis of $\mathbb{R}^{n}$ formed by the columns of $P$, then $D$ is the $B$-matrix of the linear transformation $T(\mathbf{x})=A \mathbf{x}$.
Proof
Let $\mathbf{b}_{1}, \mathbf{b}_{2}, \ldots, \mathbf{b}_{n}$ be the columns of $P$ so that $B=\left\{\mathbf{b}_{1}, \mathbf{b}_{2}, \ldots, \mathbf{b}_{n}\right\}$ is a basis. We know that for any basis in $\mathbb{R}^{n}$

$$
\mathbf{x}=P[\mathbf{x}]_{B} \Rightarrow[\mathbf{x}]_{B}=P^{-1} \mathbf{x}
$$

Let $[T]_{B}$ be the transformation matrix in the basis $B$. We know that by definition

$$
\begin{aligned}
& {[T]_{B}=\left(\left[\begin{array}{llll}
\left.T\left(\mathbf{b}_{1}\right)\right]_{B} & {\left[T\left(\mathbf{b}_{2}\right)\right]_{B}} & \ldots & \left.\left[T\left(\mathbf{b}_{n}\right)\right]_{B}\right) \quad(T(\mathbf{x})=A \mathbf{x})
\end{array}\right.\right.} \\
& =\left(\begin{array}{llll}
{\left[A \mathbf{b}_{1}\right]_{B}} & {\left[\begin{array}{lll}
\left.A \mathbf{b}_{2}\right]_{B} & \cdots & {\left[A \mathbf{b}_{n}\right.}
\end{array}\right]_{B}}
\end{array}\right) \\
& =\left(\begin{array}{llll}
P^{-1} A \mathbf{b}_{1} & P^{-1} A \mathbf{b}_{2} & \ldots & P^{-1} A \mathbf{b}_{n}
\end{array}\right) \\
& =P^{-1} A\left(\begin{array}{llll}
\mathbf{b}_{1} & \mathbf{b}_{2} & \ldots & \mathbf{b}_{n}
\end{array}\right) \\
& =P^{-1} A P=D
\end{aligned}
$$

## Transformations from $\mathbb{R}^{n}$ into $\mathbb{R}^{n}$

## Example

Let $T(\mathbf{x})=\left(\begin{array}{cc}7 & 2 \\ -4 & 1\end{array}\right) \mathbf{x}$. Find a basis $B$ in which the $B$-matrix of $T$ is diagonal. Solution We diagonalize $A$ as $A=P D P^{-1}$, with $P=\left(\begin{array}{cc}1 & 1 \\ -1 & -2\end{array}\right)$ and $D=\left(\begin{array}{ll}5 & 0 \\ 0 & 3\end{array}\right)$. We may change vectors $\mathbf{x}$ to the basis $B=\{(1,-1),(1,-2)\}$ by applying

$$
\mathbf{u}=P^{-1} \mathbf{x}
$$

Then, in this new basis, $T$ can be applied as

$$
T(\mathbf{u})=D \mathbf{u}=D P^{-1} \mathbf{x}
$$

If we now, come back to the original basis

$$
T(\mathbf{x})=P T(\mathbf{u})=P D P^{-1} \mathbf{x}=A \mathbf{x}
$$

Understanding $D$ as the transformation matrix in some basis gives us insight on its effect (in this example, an anisotropic dilation).

## Similar matrices

## Definition 4.3 (Similar matrices)

$A$ and $C$ are similar matrices iff there exists another matrix $P$ such that $A=P C P^{-1}$. Given the transformation $T(\mathbf{x})=A \mathbf{x}, C$ is the $B$-matrix of the transformation $T$, when $B$ is the basis defined by the columns of the matrix $P$.

Conversely, if $B$ is any basis and $P$ is the matrix formed by the vectors in the basis $B$, then the $B$-matrix of the transformation $T$ is $P^{-1} A P$.


## Similar matrices

## Example

Let $A=\left(\begin{array}{cc}4 & -9 \\ 4 & 8\end{array}\right), T(\mathbf{x})=A \mathbf{x}$ and $\mathbf{b}_{1}=(3,2), \mathbf{b}_{2}=(2,1)$. $A$ is not diagonalizable but the basis $B=\left\{\mathbf{b}_{1}, \mathbf{b}_{2}\right\}$ has the property that $[T]_{B}$ is triangular (it is said to be in Jordan form). According to the previous definition, the $B$-matrix of the transformation $T$ is

$$
[T]_{B}=P^{-1} A P=\left(\begin{array}{cc}
-1 & 2 \\
2 & -3
\end{array}\right)\left(\begin{array}{cc}
4 & -9 \\
4 & 8
\end{array}\right)\left(\begin{array}{ll}
3 & 2 \\
2 & 1
\end{array}\right)=\left(\begin{array}{cc}
-2 & 1 \\
0 & -2
\end{array}\right)
$$

## Numerical note

An easy way to compute $P^{-1} A P$ once we have $A P$ is to find a row equivalent matrix

$$
(P \mid A P) \sim\left(I \mid P^{-1} A P\right)
$$

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 5, Section 4:

- 5.4.1
- 5.4.3
- 5.4 .5
- 5.4.13
- 5.4 .18
- 5.4.22
- 5.4.23
- 5.4 .25
- 5.4.27 (computer)


## Outline

(6) Eigenvalues and eigenvectors

- Definition (a)
- Characteristic equation (a)
- Diagonalization (b)
- Eigenvectors and linear transformations (b)
- Complex eigenvalues (c)


## Complex eigenvalues

Complex eigenvalues are always related to a rotation around a certain axis.

## Example

Consider the linear transformation $T(\mathbf{x})=\left(\begin{array}{cc}0 & -1 \\ 1 & 0\end{array}\right) \mathbf{x}$ is a rotation of $90^{\circ}$.


Obviously, there cannot be any real eigenvector since all the vectors are rotating. All eigenvalues are complex:

$$
|A-\lambda I|=0=\lambda^{2}+1=(\lambda-i)(\lambda+i)
$$

## Complex eigenvalues

## Example (continued)

Let's see what happens if we allow applying the transformation on complex vectors:

$$
\begin{aligned}
& \left(\begin{array}{cc}
0 & -1 \\
1 & 0
\end{array}\right)\binom{1}{-i}=i\binom{1}{-i} \\
& \left(\begin{array}{cc}
0 & -1 \\
1 & 0
\end{array}\right)\binom{1}{i}=-i\binom{1}{i}
\end{aligned}
$$

## Complex eigenvalues

## Example

Find the eigenvalues and eigenvectors of $A=\left(\begin{array}{cc}\frac{1}{2} & -\frac{3}{5} \\ \frac{3}{4} & \frac{11}{10}\end{array}\right)$.
Solution
To find the eigenvalues we solve the characteristic equation:

$$
0=|A-\lambda I|=\left|\begin{array}{cc}
\frac{1}{2}-\lambda & -\frac{3}{5} \\
\frac{3}{4} & \frac{11}{10}-\lambda
\end{array}\right|=\lambda^{2}-\frac{8}{5} \lambda+1 \Rightarrow \lambda=\frac{4}{5} \pm \frac{3}{5} i
$$

MATLAB: $A=[1 / 2-3 / 5 ; 3 / 411 / 10] ; 1=e i g s(A)$

## Complex eigenvalues

## Example (continued)

$$
\lambda_{1}=\frac{4}{5}-\frac{3}{5} i
$$

$$
\left.\begin{array}{rl}
A-\lambda_{1} I & =\left(\begin{array}{cc}
\frac{1}{2}-\left(\frac{4}{5}-\frac{3}{5} i\right) & -\frac{3}{5} \\
\frac{3}{4} & \frac{11}{10}-\left(\frac{4}{5}-\frac{3}{5} i\right)
\end{array}\right)=\left(\begin{array}{cc}
-\frac{3}{10}+\frac{3}{5} i & -\frac{3}{5} \\
\frac{3}{4} & \frac{3}{10}+\frac{3}{5} i
\end{array}\right) \\
1 & \frac{2}{5}+\frac{4}{5} i \\
0 & 0
\end{array}\right) \Rightarrow x_{1}=-\left(\frac{2}{5}+\frac{4}{5} i\right) x_{2} \Rightarrow \mathbf{v}_{1}=\binom{-2-4 i}{5}
$$

MATLAB:

$$
\begin{aligned}
& A_{-} l \mathrm{I}=A_{-1}(1) * \operatorname{eye}(2) ; \\
& A_{-} l I(1,:)=A_{-} l I(1,:) / A_{-} l I(1,1) \\
& A_{-} l I(2,:)=A_{-} l I(2,:)-A_{-} l I(1,:) * A_{-} l I(2,1) \\
& \lambda_{2}=\frac{4}{5}+\frac{3}{5} i=\lambda_{1}^{*}
\end{aligned}
$$

$$
A-\lambda_{2} I \sim\left(\begin{array}{cc}
1 & \frac{2}{5}-\frac{4}{5} i \\
0 & 0
\end{array}\right) \Rightarrow x_{1}=-\left(\frac{2}{5}-\frac{4}{5} i\right) x_{2} \Rightarrow \mathbf{v}_{2}=\binom{-2+4 i}{5}=\mathbf{v}_{1}^{*}
$$

## Complex eigenvalues

## Example (continued)

The application of $A$ on $\mathbb{R}^{2}$ is a rotation. To see this, we may start with $\mathbf{x}_{0}=(2,0)$ and calculate

$$
\begin{aligned}
& \mathbf{x}_{1}=A \mathbf{x}_{0}=\left[\begin{array}{lr}
.5 & -.6 \\
.75 & 1.1
\end{array}\right]\left[\begin{array}{l}
2 \\
0
\end{array}\right]=\left[\begin{array}{l}
1.0 \\
1.5
\end{array}\right] \\
& \mathbf{x}_{2}=A \mathbf{x}_{1}=\left[\begin{array}{lr}
.5 & -.6 \\
.75 & 1.1
\end{array}\right]\left[\begin{array}{l}
1.0 \\
1.5
\end{array}\right]=\left[\begin{array}{r}
-.4 \\
2.4
\end{array}\right] \\
& \mathbf{x}_{3}=A \mathbf{x}_{2}, \ldots
\end{aligned}
$$

Figure 1 shows $\mathbf{x}_{0}, \ldots, \mathbf{x}_{8}$ as larger dots. The smaller dots are the locations of $\mathbf{x}_{9}, \ldots$, $\mathbf{x}_{100}$. The sequence lies along an elliptical orbit.


## Complex eigenvalues

## Definition 5.1 (Conjugate of a vector and matrix)

The conjugate of a vector is defined as

$$
\mathbf{v}=\left(\begin{array}{c}
v_{1} \\
v_{2} \\
\ldots \\
v_{n}
\end{array}\right) \Rightarrow \mathbf{v}^{*}=\left(\begin{array}{c}
v_{1}^{*} \\
v_{2}^{*} \\
\ldots \\
v_{n}^{*}
\end{array}\right)
$$

In the same way, the conjugate of a matrix is defined as

$$
A=\left(\begin{array}{cccc}
a_{11} & a_{12} & \ldots & a_{1 n} \\
a_{21} & a_{22} & \ldots & a_{2 n} \\
\ldots & \ldots & \ldots & \ldots \\
a_{m 1} & a_{m 2} & \ldots & a_{m n}
\end{array}\right) \Rightarrow A^{*}=\left(\begin{array}{cccc}
a_{11}^{*} & a_{12}^{*} & \ldots & a_{1 n}^{*} \\
a_{21}^{*} & a_{22}^{*} & \ldots & a_{2 n}^{*} \\
\ldots & \ldots & \ldots & \ldots \\
a_{m 1}^{*} & a_{m 2}^{*} & \ldots & a_{m n}^{*}
\end{array}\right)
$$

Theorem 5.1 (Properties)

$$
\begin{array}{ll}
(r \mathbf{v})^{*}=r^{*} \mathbf{v}^{*} & (A B)^{*}=A^{*} B^{*} \\
(A \mathbf{v})^{*}=A^{*} \mathbf{v}^{*} & (r A)^{*}=r^{*} A^{*}
\end{array}
$$

## Eigenanalysis of a real matrix that acts on $\mathbb{C}^{n}$

## Theorem 5.2

Let $A \in \mathcal{M}_{n \times n}$ be a matrix with real coefficients. If $\lambda$ is an eigenvalue of $A$, then $\lambda^{*}$ is also an eigenvalue. If $\mathbf{v}$ is an eigenvector associated to $\lambda$, then $\mathbf{v}^{*}$ is an eigenvector associated to $\lambda^{*}$.
Proof
If $\lambda$ is an eigenvalue and $\mathbf{v}$ one of its eigenvectors, then we know that

$$
A \mathbf{v}=\lambda \mathbf{v}
$$

If we now conjugate both sides

$$
(A \mathbf{v})^{*}=(\lambda \mathbf{v})^{*} \Rightarrow A \mathbf{v}^{*}=\lambda^{*} \mathbf{v}^{*}
$$

(Remind that $A$ has real coefficients and that's why $A^{*}=A$ ).
The previous equation means that $\mathbf{v}^{*}$ is also an eigenvector of $A$ and that $\lambda^{*}$ is its eigenvalue.

## Eigenanalysis of a real matrix that acts on $\mathbb{C}^{n}$

## Example

Let $A=\left(\begin{array}{cc}a & -b \\ b & a\end{array}\right)$. Its eigenvalues are $\lambda=a \pm b i$ and the corresponding eigenvectors $\mathbf{v}=\binom{1}{ \pm i}$.

$$
\begin{aligned}
\left(\begin{array}{cc}
a & -b \\
b & a
\end{array}\right)\binom{1}{-i} & =\binom{a+b i}{b-a i}=(a+b i)\binom{1}{-i} \\
\left(\begin{array}{cc}
a & -b \\
b & a
\end{array}\right)\binom{1}{i} & =\binom{a-b i}{b+a i}=(a-b i)\binom{1}{i}
\end{aligned}
$$

In particular if $a=\cos (\phi)$ and $b=\sin (\phi)$, then we have a rotation matrix whose eigenvalues are

$$
\left(\begin{array}{cc}
\cos (\phi) & -\sin (\phi) \\
\sin (\phi) & \cos (\phi)
\end{array}\right) \Rightarrow \lambda=\cos (\phi) \pm \sin (\phi) i=e^{ \pm i \phi}
$$

## Eigenanalysis of a real matrix that acts on $\mathbb{C}^{n}$

## Example on Slide 60 (continued)

Let $A=\left(\begin{array}{cc}\frac{1}{2} & -\frac{3}{5} \\ \frac{3}{4} & \frac{11}{10}\end{array}\right)$. Consider $\lambda_{1}=\frac{4}{5}-\frac{3}{5} i$ and its corresponding eigenvector $\mathbf{v}_{1}=(-2-4 i, 5)$. Now, we construct the matrix

$$
P=\left(\operatorname{Re}\left\{\mathbf{v}_{1}\right\} \quad \operatorname{Im}\left\{\mathbf{v}_{1}\right\}\right)=\left(\begin{array}{cc}
-2 & -4 \\
5 & 0
\end{array}\right)
$$

and make a change of basis to the basis whose vectors are the columns of $P$ :

$$
C=P^{-1} A P=\left(\begin{array}{cc}
\frac{4}{5} & -\frac{3}{5} \\
\frac{3}{5} & \frac{4}{5}
\end{array}\right)=\left(\begin{array}{cc}
\cos \left(36.87^{\circ}\right) & -\sin \left(36.87^{\circ}\right) \\
\sin \left(36.87^{\circ}\right) & \cos \left(36.87^{\circ}\right)
\end{array}\right)
$$

That is, $C$ is a pure rotation and thanks to the change of basis we obtain an elliptical rotation as shown in Slide 62.

## Eigenanalysis of a real matrix that acts on $\mathbb{C}^{n}$

## Theorem 5.3

Let $A$ be a real, $2 \times 2$ matrix with complex eigenvalue $\lambda=a-b i(b \neq 0)$ and an associated eigenvector in $\mathbb{C}^{2}$. Then

$$
A=P C P^{-1}
$$

where

$$
P=(\operatorname{Re}\{\mathbf{v}\} \quad \operatorname{Im}\{\mathbf{v}\})
$$

and

$$
C=\left(\begin{array}{cc}
a & -b \\
b & a
\end{array}\right)
$$

Proof
It makes use of

$$
\begin{aligned}
& \operatorname{Re}\{A \mathbf{v}\}=A \operatorname{Re}\{\mathbf{v}\} \\
& \operatorname{Im}\{A \mathbf{v}\}=A \operatorname{Im}\{\mathbf{v}\}
\end{aligned}
$$

## Eigenanalysis of a real matrix that acts on $\mathbb{C}^{n}$

## Example: Rotations extend to higher dimensions

Consider $A=\left(\begin{array}{ccc}4 & -\frac{3}{5} & 0 \\ \frac{5}{5} & \frac{4}{5} & 0 \\ 0 & 0 & 1.07\end{array}\right)$. This is the rotation previously described in the
$X Y$ plane plus a scaling in the $Z$ direction. Any point in the $X Y$ (for instance, $\left.\mathbf{w}_{0}=(2,0,0)\right)$ plane rotates within the plane. Any point outside the plane (for instance, $\mathbf{x}_{0}=(2,0,1)$ rotates in $X Y$ and shifts along $\left.Z\right)$. The following figure shows the successive application of $A$ on $\mathbf{w}_{0}$ and $\mathbf{x}_{0}$.


## Exercises

## Exercises

From Lay (3rd ed.), Chapter 5, Section 5:

- 5.5.1
- 5.5.7
- 5.5.13
- 5.5.23
- 5.5.24
- 5.5.25
- 5.5.26
- 5.5.27


## Outline

(6) Eigenvalues and eigenvectors

- Definition (a)
- Characteristic equation (a)
- Diagonalization (b)
- Eigenvectors and linear transformations (b)
- Complex eigenvalues (c)


# Chapter 7. Orthogonality and least squares 

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Biomedical Engineering

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## Outline

(7) Orthogonality and least squares

- Inner product, length and orthogonality (a)
- Orthogonal sets, bases and matrices (a)
- Orthogonal projections (b)
- Gram-Schmidt orthogonalization (b)
- Least squares (c)
- Least-squares linear regression (c)
- Inner product spaces (d)
- Applications of inner product spaces (d)


## References


D. Lay. Linear algebra and its applications (3rd ed). Pearson (2006). Chapter 6.

## A little bit of history

Least squares was first used to solve problems in geodesy (Andrien-Marie Legendre, 1805) and astronomy (Carl Friedrich Gauss, 1809). Gauss made the connection of this method to the distribution of measurement errors. Currently it is one of the best understood and most widely spread methods.


## Applications

In this example Least Squares are used to plan a radiation therapy.


Bedford, J. L. Sinogram analysis of aperture optimization by iterative least-squares in volumetric modulated arc therapy. Physics in Medicine and Biology, 2013, 58, 1235-1250

## Applications

Traditionally, control applications were formulated in a least-squares setup. Currently, they have found more sophisticated goal functions that can be regarded as evolved versions of least squares.


## Outline

(7) Orthogonality and least squares

- Inner product, length and orthogonality (a)
- Orthogonal sets, bases and matrices (a)
- Orthogonal projections (b)
- Gram-Schmidt orthogonalization (b)
- Least squares (c)
- Least-squares linear regression (c)
- Inner product spaces (d)
- Applications of inner product spaces (d)


## Inner product

## Definition 1.1 (Inner product or dot product)

Let $\mathbf{u}, \mathbf{v} \in \mathbb{R}^{n}$ be two vectors. The inner product or dot product between these two vectors is defined as

$$
\mathbf{u} \cdot \mathbf{v}=\langle\mathbf{u}, \mathbf{v}\rangle \triangleq \sum_{i=1}^{n} u_{i} v_{i}
$$

## Theorem 1.1

If we considered $\mathbf{u}$ and $\mathbf{v}$ to be column vectors $\left(\in \mathcal{M}_{n \times 1}\right)$, then

$$
\mathbf{u} \cdot \mathbf{v}=\mathbf{u}^{T} \mathbf{v}
$$

## Example

Let $\mathbf{u}=(2,-5,-1)$ and $\mathbf{v}=(3,2,-3)$.

$$
\mathbf{u} \cdot \mathbf{v}=2 \cdot 3+(-5) \cdot 2+1 \cdot(-3)=-1
$$

## Inner product

## Theorem 1.2

For any three vectors $\mathbf{u}, \mathbf{v}, \mathbf{w} \in \mathbb{R}^{n}$ and any scalar $r \in \mathbb{R}$ it is verified that
(1) $\mathbf{u} \cdot \mathbf{v}=\mathbf{v} \cdot \mathbf{u}$
(2) $(\mathbf{u}+\mathbf{v}) \cdot \mathbf{w}=\mathbf{u} \cdot \mathbf{w}+\mathbf{v} \cdot \mathbf{w}$
(3) $(r \mathbf{u}) \cdot \mathbf{v}=r(\mathbf{u} \cdot \mathbf{v})=\mathbf{u} \cdot(r \mathbf{v})$
(4) $\mathbf{u} \cdot \mathbf{u} \geq 0$
(1) $\mathbf{u} \cdot \mathbf{u}=0 \Leftrightarrow \mathbf{u}=\mathbf{0}$

## Corollary

$$
\left(r_{1} \mathbf{u}_{1}+r_{2} \mathbf{u}_{2}+\ldots+r_{p} \mathbf{u}_{p}\right) \cdot \mathbf{v}=r_{1}\left(\mathbf{u}_{1} \cdot \mathbf{v}\right)+r_{2}\left(\mathbf{u}_{2} \cdot \mathbf{v}\right)+\ldots+r_{p}\left(\mathbf{u}_{p} \cdot \mathbf{v}\right)
$$

## Length

Definition 1.2 (Length of a vector)
Given any vector $\mathbf{v}$, its length is defined as

$$
\|\mathbf{v}\| \triangleq \sqrt{\mathbf{v} \cdot \mathbf{v}}
$$

## Theorem 1.3

Given any vector $\mathbf{v} \in \mathbb{R}^{n}$

$$
\|\mathbf{v}\|=\sqrt{v_{1}^{2}+v_{2}^{2}+\ldots+v_{n}^{2}}
$$

## Example

The length of $\mathbf{v}=(1,-2,2,0)$ is

$$
\|\mathbf{v}\|=\sqrt{1^{2}+(-2)^{2}+2^{2}+0^{2}}=3
$$

## Length

## Theorem 1.4

For any vector $\mathbf{v}$ and any scalar $r$ it is verified that

$$
\|r \mathbf{v}\|=\mid r\|\mathbf{v}\|
$$

## Proof

It will be given only for $\mathbf{v} \in \mathbb{R}^{n}$ :

$$
\begin{aligned}
\|r \mathbf{v}\| & =\sqrt{\left(r v_{1}\right)^{2}+\left(r v_{2}\right)^{2}+\ldots+\left(r v_{n}\right)^{2}}=\sqrt{r^{2}\left(v_{1}^{2}+v_{2}^{2}+\ldots+v_{n}^{2}\right)} \quad \text { (q.e.d.) } \\
& =\sqrt{r^{2}} \sqrt{v_{1}^{2}+v_{2}^{2}+\ldots+v_{n}^{2}}=\mid r\|\mathbf{v}\|
\end{aligned}
$$

## Example (continued)

Find a vector of unit length that has the same direction as $\mathbf{v}=(1,-2,2,0)$. Solution

$$
\mathbf{u}_{\mathbf{v}}=\frac{v}{\|\mathbf{v}\|}=\left(\frac{1}{3},-\frac{2}{3}, \frac{2}{3}, 0\right) \Rightarrow\left\|\mathbf{u}_{\mathbf{v}}\right\|=\sqrt{\frac{1}{9}+\frac{4}{9}+\frac{4}{9}+0}=1
$$

## Distance

## Definition 1.3 (Distance in $\mathbb{R}$ )

The distance between any two numbers $a, b \in \mathbb{R}$ can be defined as

$$
d(a, b)=|a-b|
$$

## Example

Calculate the distance between 2 and 8 as well as between -3 and 4 .

$|2-8|=|-6|=6$ or $|8-2|=|6|=6$


$$
|(-3)-4|=|-7|=7 \text { or }|4-(-3)|=|7|=7
$$

FIGURE 3 Distances in $\mathbb{R}$.

## Distance

## Definition 1.4 (Distance in $\mathbb{R}^{n}$ )

The distance between any two vectors $\mathbf{u}, \mathbf{v} \in \mathbb{R}^{n}$ can be defined as

$$
d(\mathbf{u}, \mathbf{v})=\|\mathbf{u}-\mathbf{v}\|
$$

## Example

Calculate the distance between $\mathbf{u}=(7,1)$ and $\mathbf{v}=(3,2)$

$$
d(\mathbf{u}, \mathbf{v})=\|(7,1)-(3,2)\|=\|(4,-1)\|=\sqrt{4^{2}+1^{2}}=\sqrt{17}
$$



FIGURE 4 The distance between $\mathbf{u}$ and $\mathbf{v}$ is the length of $\mathbf{u}-\mathbf{v}$.

## Distance

## Example

For any two vectors in $\mathbb{R}^{3}, \mathbf{u}$ and $\mathbf{v}$, the distance can be calculated through

$$
\begin{gathered}
d(\mathbf{u}, \mathbf{v})=\|\mathbf{u}-\mathbf{v}\|=\left\|\left(u_{1}-v_{1}, u_{2}-v_{2}, u_{3}-v_{3}\right)\right\|= \\
\sqrt{\left(u_{1}-v_{1}\right)^{2}+\left(u_{2}-v_{2}\right)^{2}+\left(u_{3}-v_{3}\right)^{2}}
\end{gathered}
$$

## Orthogonality

## Example

Any two vectors in $\mathbb{R}^{2}, \mathbf{u}$ and $\mathbf{v}$, are orthogonal if $d(\mathbf{u}, \mathbf{v})=d(\mathbf{u},-\mathbf{v})$


$$
\begin{gathered}
d^{2}(\mathbf{u}, \mathbf{v})=\|\mathbf{u}-\mathbf{v}\|^{2}=(\mathbf{u}-\mathbf{v}) \cdot(\mathbf{u}-\mathbf{v})=\mathbf{u} \cdot \mathbf{u}+\mathbf{v} \cdot \mathbf{v}-2 \mathbf{u} \cdot \mathbf{v}=\|\mathbf{u}\|^{2}+\|\mathbf{v}\|^{2}-2 \mathbf{u} \cdot \mathbf{v} \\
d^{2}(\mathbf{u},-\mathbf{v})=\|\mathbf{u}+\mathbf{v}\|^{2}=(\mathbf{u}+\mathbf{v}) \cdot(\mathbf{u}+\mathbf{v})=\mathbf{u} \cdot \mathbf{u}+\mathbf{v} \cdot \mathbf{v}+2 \mathbf{u} \cdot \mathbf{v}=\|\mathbf{u}\|^{2}+\|\mathbf{v}\|^{2}+2 \mathbf{u} \cdot \mathbf{v} \\
d^{2}(\mathbf{u}, \mathbf{v})=d^{2}(\mathbf{u},-\mathbf{v}) \Rightarrow-2 \mathbf{u} \cdot \mathbf{v}=2 \mathbf{u} \cdot \mathbf{v} \Rightarrow \mathbf{u} \cdot \mathbf{v}=0
\end{gathered}
$$

## Orthogonality

Definition 1.5 (Orthogonality between two vectors)
Any two different vectors, $\mathbf{u}$ and $\mathbf{v}$, in a vector space $V$ are orthogonal iff

$$
\mathbf{u} \cdot \mathbf{v}=0
$$

## Corollary

$\mathbf{0}$ is orthogonal to any other vector.

## Theorem 1.5 (Pythagorean theorem)

Any two vectors, $\mathbf{u}$ and $\mathbf{v}$, in a vector space $V$ are orthogonal iff

$$
\|\mathbf{u}+\mathbf{v}\|^{2}=\|\mathbf{u}\|^{2}+\|\mathbf{v}\|^{2}
$$

## Orthogonality

## Definition 1.6 (Orthogonality between vector and vector space)

Let $\mathbf{u}$ be a vector in a vector space $V$ and $W$ a vector subspace of $V$. $\mathbf{u}$ is orthogonal to $W$ if $\mathbf{u}$ is orthogonal to all vectors in $W$. The set of all vectors orthogonal to $W$ is denoted as $W^{\perp}$ (the orthogonal complement of $W$ ).

## Example

Let $W$ be a plane in $\mathbb{R}^{3}$ passing through the origin and $L$ be a line, passing through the origin and perpendicular to $W$. For any vector $\mathbf{w} \in W$ and any vector $z \in L$ we have

$$
\mathbf{w} \cdot \mathbf{z}=0
$$

Therefore,

$$
L=W^{\perp} \Leftrightarrow W=L^{\perp}
$$



## Orthogonality

## Theorem 1.6

Let $W$ be a vector subspace of a vector space $V$.
(1) $\mathbf{x} \in W^{\perp}$ iff $\mathbf{x}$ is orthogonal to every vector in a set that spans $W$.
(2) $W^{\perp}$ is a vector subspace of $V$.

## Theorem 1.7

Let $A \in \mathcal{M}_{m \times n}$, then
(1) $(\operatorname{Row}\{A\})^{\perp}=\operatorname{Nul}\{A\}$
(2) $(\operatorname{Col}\{A\})^{\perp}=\operatorname{Nul}\left\{A^{T}\right\}$

FIGURE 8 The fundamental subspaces determined by an $m \times n$ matrix $A$.

## Orthogonality

## Proof $\operatorname{Nul}\{A\} \subseteq(\operatorname{Row}\{A\})^{\perp}$

Consider the rows of $A, \mathbf{a}_{i}(i=1,2, \ldots, m)$ as column vectors, then for any vector $\mathbf{x} \in \operatorname{Nul}\{A\}$ we know

$$
A \mathbf{x}=\mathbf{0} \Rightarrow\left(\begin{array}{c}
\mathbf{a}_{1}^{T} \\
\mathbf{a}_{2}^{T} \\
\ldots \\
\mathbf{a}_{m}^{T}
\end{array}\right) \mathbf{x}=\left(\begin{array}{c}
\mathbf{a}_{1}^{T} \mathbf{x} \\
\mathbf{a}_{2}^{T} \mathbf{x} \\
\ldots \dddot{ } \\
\mathbf{a}_{m}^{T} \mathbf{x}
\end{array}\right)=\left(\begin{array}{c}
\mathbf{a}_{1} \cdot \mathbf{x} \\
\mathbf{a}_{2} \cdot \mathbf{x} \\
\ldots \\
\mathbf{a}_{m} \cdot \mathbf{x}
\end{array}\right)=\left(\begin{array}{c}
0 \\
0 \\
\ldots \\
0
\end{array}\right)
$$

Consequently, $\mathbf{x}$ is orthogonal to all the rows of $A$, which $\operatorname{span} \operatorname{Row}\{A\}$ and by the previous theorem, $\mathbf{x} \in(\operatorname{Row}\{A\})^{\perp}$
Proof $\operatorname{Nul}\{A\} \supseteq(\operatorname{Row}\{A\})^{\perp}$
Conversely, let $\mathbf{x} \in(\operatorname{Row}\{A\})^{\perp}$, then by the previous theorem we know that

$$
\mathbf{a}_{i} \cdot \mathbf{x} \text { for } i=1,2, \ldots, m \Rightarrow A \mathbf{x}=\mathbf{0}
$$

So, $\mathbf{x} \in \operatorname{Nul}\{A\}$

## Orthogonality

$\operatorname{Proof}(\operatorname{Col}\{A\})^{\perp}=\operatorname{Nul}\left\{A^{T}\right\}$
Let's define $B=A^{T}$. By the first part of this theorem, we know

$$
(\operatorname{Row}\{B\})^{\perp}=\operatorname{Nul}\{B\} \Rightarrow\left(\operatorname{Row}\left\{A^{T}\right\}\right)^{\perp}=\operatorname{Nul}\left\{A^{T}\right\} \Rightarrow(\operatorname{Col}\{A\})^{\perp}=\operatorname{Nul}\left\{A^{T}\right\}
$$

## Theorem 1.8

For any two vectors $\mathbf{u}$ and $\mathbf{v}$ in a vector space $V$, the angle between the two can be measured through the dot product:

$$
\mathbf{u} \cdot \mathbf{v}=\|\mathbf{u}\|\|\mathbf{v}\| \cos \theta
$$

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 6, Section 1:

- 6.1.15
- 6.1.22
- 6.1.24
- 6.1.26
- 6.1.28
- 6.1.30
- 6.1.32 (computer)


## Outline

(7) Orthogonality and least squares

- Inner product, length and orthogonality (a)
- Orthogonal sets, bases and matrices (a)
- Orthogonal projections (b)
- Gram-Schmidt orthogonalization (b)
- Least squares (c)
- Least-squares linear regression (c)
- Inner product spaces (d)
- Applications of inner product spaces (d)


## Orthogonal sets

## Definition 2.1 (Orthogonal set)

Let $S=\left\{\mathbf{u}_{1}, \mathbf{u}_{2}, \ldots, \mathbf{u}_{p}\right\}$ be a set of vectors. $S$ is an orthogonal set iff

$$
\mathbf{u}_{i} \cdot \mathbf{u}_{j}=0 \quad \forall i, j \in\{1,2, \ldots, p\} i \neq j
$$

## Example

Let $\mathbf{u}_{1}=(3,1,1), \mathbf{u}_{2}=(-1,2,1), \mathbf{u}_{3}=\left(-\frac{1}{2},-2, \frac{7}{2}\right)$. Check whether the set $S=\left\{\mathbf{u}_{1}, \mathbf{u}_{2}, \mathbf{u}_{3}\right\}$ is orthogonal.
Solution

$$
\begin{aligned}
& \mathbf{u}_{1} \cdot \mathbf{u}_{2}=3 \cdot(-1)+1 \cdot 2+1 \cdot 1=0 \\
& \mathbf{u}_{1} \cdot \mathbf{u}_{3}=3 \cdot\left(-\frac{1}{2}\right)+1 \cdot(-2)+1 \cdot\left(\frac{7}{2}\right)=0 \\
& \mathbf{u}_{2} \cdot \mathbf{u}_{3}=(-1) \cdot\left(-\frac{1}{2}\right)+2 \cdot(-2)+1 \cdot\left(\frac{7}{2}\right)=0
\end{aligned}
$$

## Orthogonal sets

## Theorem 2.1

If $S$ is an orthogonal set of non-null vectors, then $S$ is linearly independent and, consequently, it is a basis of the subspace spanned by $S$.
Proof
Let $\mathbf{u}_{i}(i=1,2, \ldots, p)$ be the elements of $S$. Let us assume that $S$ is linearly dependent. Then, there exists coefficients $c_{1}, c_{2}, \ldots, c_{p}$ not all of them null such that

$$
\mathbf{0}=c_{1} \mathbf{u}_{1}+c_{2} \mathbf{u}_{2}+\ldots+c_{p} \mathbf{u}_{p}
$$

Now, we compute the inner product with $\mathbf{u}_{1}$

$$
\begin{gathered}
\mathbf{0} \cdot \mathbf{u}_{1}=\left(c_{1} \mathbf{u}_{1}+c_{2} \mathbf{u}_{2}+\ldots+c_{p} \mathbf{u}_{p}\right) \cdot \mathbf{u}_{1} \\
0=c_{1}\left(\mathbf{u}_{1} \cdot \mathbf{u}_{1}\right)+c_{2}\left(\mathbf{u}_{2} \cdot \mathbf{u}_{1}\right)+\ldots+c_{p}\left(\mathbf{u}_{p} \cdot \mathbf{u}_{1}\right)=c_{1}\left\|\mathbf{u}_{1}\right\|^{2} \Rightarrow c_{1}=0
\end{gathered}
$$

Multiplying by $\mathbf{u}_{i}(i=2,3, \ldots, p)$ we can show that all $c_{i}$ 's are 0 , and, therefore, the set $S$ is linearly independent.

## Orthogonal basis

## Definition 2.2 (Orthogonal basis)

$A$ set of vectors $B$ is an ortohogonal basis of a vector space $V$ if it is an ortohogonal set and it is a basis of $V$.

## Theorem 2.2

Let $\left\{\mathbf{u}_{1}, \mathbf{u}_{2}, \ldots, \mathbf{u}_{p}\right\}$ be an orthogonal basis for a vector space $V$, for each $\mathbf{x} \in V$ we have

$$
\mathbf{x}=\frac{\mathbf{x} \cdot \mathbf{u}_{1}}{\left\|\mathbf{u}_{1}\right\|^{2}} \mathbf{u}_{1}+\frac{\mathrm{x} \cdot \mathbf{u}_{2}}{\left\|\mathbf{u}_{2}\right\|^{2}} \mathbf{u}_{2}+\ldots+\frac{\mathrm{x} \cdot \mathbf{u}_{\mathrm{p}}}{\left\|\mathbf{u}_{p}\right\|^{2}} \mathbf{u}_{p}
$$

Proof
If $\mathbf{x}$ is in $V$, then it can be expressed as a linear combination of the vectors in a basis of $V$

$$
\mathbf{x}=c_{1} \mathbf{u}_{1}+c_{2} \mathbf{u}_{2}+\ldots+c_{p} \mathbf{u}_{p}
$$

Now, we calculate the dot product with $\mathbf{u}_{1}$

$$
\mathbf{x} \cdot \mathbf{u}_{1}=\left(c_{1} \mathbf{u}_{1}+c_{2} \mathbf{u}_{2}+\ldots+c_{p} \mathbf{u}_{p}\right) \cdot \mathbf{u}_{1}=c_{1}\left\|\mathbf{u}_{1}\right\|^{2} \Rightarrow c_{1}=\frac{x \cdot \mathbf{u}_{1}}{\left\|\mathbf{u}_{1}\right\|^{2}}
$$

## Orthogonal basis

## Example

Let $\mathbf{u}_{1}=(3,1,1), \mathbf{u}_{2}=(-1,2,1), \mathbf{u}_{3}=\left(-\frac{1}{2},-2, \frac{7}{2}\right)$, and $B=\left\{\mathbf{u}_{1}, \mathbf{u}_{2}, \mathbf{u}_{3}\right\}$ be an orthogonal basis of $\mathbb{R}^{3}$. Let $\mathbf{x}=(6,1,-8)$. The coordinates of $\mathbf{x}$ in $B$ are given by

$$
\begin{aligned}
& \mathbf{x} \cdot \mathbf{u}_{1}=11 \quad \mathbf{x} \cdot \mathbf{u}_{2}=-12 \quad \mathbf{x} \cdot \mathbf{u}_{1}=-33 \\
& \left\|\mathbf{u}_{1}\right\|^{2}=11 \quad\left\|\mathbf{u}_{2}\right\|^{2}=6 \quad\left\|\mathbf{u}_{3}\right\|^{2}=\frac{33}{2} \\
& \mathbf{x}=\frac{11}{11} \mathbf{u}_{1}+\frac{-12}{6} \mathbf{u}_{2}+\frac{-33}{\frac{33}{2}} \mathbf{u}_{3} \\
& =\mathbf{u}_{1}-2 \mathbf{u}_{2}-2 \mathbf{u}_{3}
\end{aligned}
$$

The coordinates of $\mathbf{x}$ in the basis $B$ are

$$
[\mathbf{x}]_{B}=(1,-2,-2)
$$

## Orthogonal projections

## Orthogonal projection onto a vector

Consider a vector $\mathbf{y}$ and another one $\mathbf{u}$. Let us assume we want to decompose $\mathbf{y}$ as the sum of two orthogonal vectors $\hat{\mathbf{y}}$ (along the direction of $\mathbf{u}$ ) and another vector z (orthogonal to $\mathbf{u}$ ):

$$
\begin{aligned}
& \mathbf{y}=\hat{\mathbf{y}}+\mathbf{z}=\alpha \mathbf{u}+\mathbf{z} \Rightarrow \\
& \mathbf{z}=\mathbf{y}-\hat{\mathbf{y}}
\end{aligned}
$$



## FIGURE 2

Finding $\alpha$ to make $\mathbf{y}-\hat{\mathbf{y}}$ orthogonal to $\mathbf{u}$.
We need to find $\alpha$ that makes $\mathbf{u}$ and $\mathbf{z}$ orthogonal.

$$
0=\mathbf{z} \cdot \mathbf{u}=(\mathbf{y}-\alpha \mathbf{u}) \cdot \mathbf{u}=\mathbf{y} \cdot \mathbf{u}-\alpha\|\mathbf{u}\|^{2} \Rightarrow \alpha=\frac{\mathbf{y} \cdot \mathbf{u}}{\|\mathbf{u}\|^{2}}
$$

$\hat{\mathbf{y}}$ is the orthogonal projection of $\mathbf{y}$ onto $\mathbf{u}$.

## Orthogonal projections

## Example

Let $\mathbf{y}=(7,6)$ and $\mathbf{u}=(4,2)$. Then,

$$
\left.\begin{array}{c}
\hat{\mathbf{y}}=\frac{\mathbf{y} \cdot \mathbf{u}}{\|\mathbf{u}\|^{2}} \mathbf{u}=\frac{40}{20} \mathbf{u}=2 \mathbf{u}=\binom{8}{4} \\
\|\mathbf{u}\|^{2}=20
\end{array}\right\} \Rightarrow \begin{gathered}
\mathbf{z}=\mathbf{y}-\hat{\mathbf{y}}=\binom{7}{6}-\binom{8}{4}=\binom{-1}{2} \\
d(\mathbf{y}, \hat{\mathbf{y}})=\|\mathbf{y}-\hat{\mathbf{y}}\|=\|\mathbf{z}\|=\sqrt{(-1)^{2}+2^{2}}=\sqrt{5}
\end{gathered}
$$

## Orthonormal set

## Definition 2.3 (Orthonormal set)

$\left\{\mathbf{u}_{1}, \mathbf{u}_{2}, \ldots, \mathbf{u}_{p}\right\}$ is an orthonormal set if it is an orthogonal set and all $\mathbf{u}_{i}$ vectors have unit length.

## Example

Show that the set $\left\{\mathbf{u}_{1}, \mathbf{u}_{2}, \mathbf{u}_{3}\right\}$ is orthonormal, with

$$
\mathbf{u}_{1}=\frac{1}{\sqrt{11}}\left(\begin{array}{l}
3 \\
1 \\
1
\end{array}\right) \quad \mathbf{u}_{2}=\frac{1}{\sqrt{6}}\left(\begin{array}{c}
-1 \\
2 \\
1
\end{array}\right) \quad \mathbf{u}_{3}=\frac{1}{\sqrt{66}}\left(\begin{array}{c}
-1 \\
-4 \\
7
\end{array}\right)
$$

## Solution

Let's check that they are orthogonal:

$$
\begin{aligned}
& \mathbf{u}_{1} \cdot \mathbf{u}_{2}=\frac{1}{\sqrt{11}} \frac{1}{\sqrt{6}}(3 \cdot(-1)+1 \cdot 2+1 \cdot 1)=0 \\
& \mathbf{u}_{1} \cdot \mathbf{u}_{3}=\frac{1}{\sqrt{11}} \frac{1}{\sqrt{66}}(3 \cdot(-1)+1 \cdot(-4)+1 \cdot 7)=0 \\
& \mathbf{u}_{2} \cdot \mathbf{u}_{3}=\frac{1}{\sqrt{6}} \frac{1}{\sqrt{66}}((-1) \cdot(-1)+(2) \cdot(-4)+(1) \cdot 7)=0
\end{aligned}
$$

## Orthonormal set

## Example (continued)

Now, let's check that they have unit length:

$$
\begin{aligned}
& \left\|\mathbf{u}_{1}\right\|=\sqrt{\left(\frac{1}{\sqrt{11}}\right)^{2}\left(3^{2}+1^{2}+1^{2}\right)}=\sqrt{\frac{9+1+1}{11}}=1 \\
& \left\|\mathbf{u}_{2}\right\|=\sqrt{\left(\frac{1}{\sqrt{6}}\right)^{2}\left((-1)^{2}+2^{2}+1^{2}\right)}=\sqrt{\frac{1+4+1}{6}}=1 \\
& \left\|\mathbf{u}_{3}\right\|=\sqrt{\left(\frac{1}{\sqrt{66}}\right)^{2}\left((-1)^{2}+(-4)^{2}+7^{2}\right)}=\sqrt{\frac{1+16+49}{66}}=1
\end{aligned}
$$

## Theorem 2.3

If $S=\left\{\mathbf{u}_{1}, \mathbf{u}_{2}, \ldots, \mathbf{u}_{n}\right\}$ is an orthonormal set, then it is an orthonormal basis of Span $\{S\}$.

## Example

$\left\{\mathbf{e}_{1}, \mathbf{e}_{2}, \ldots, \mathbf{e}_{n}\right\}$ is an orthonormal basis of $\mathbb{R}^{n}$.

## Orthonormal basis

## Theorem 2.4

Let $S=\left\{\mathbf{u}_{1}, \mathbf{u}_{2}, \ldots, \mathbf{u}_{n}\right\}$ is an orthogonal set of vectors, then the set $S^{\prime}=\left\{\mathbf{u}_{1}^{\prime}, \mathbf{u}_{2}^{\prime}, \ldots, \mathbf{u}_{n}^{\prime}\right\}$ where

$$
\mathbf{u}_{i}^{\prime}=\frac{\mathbf{u}_{i}}{\left\|\mathbf{u}_{i}\right\|}
$$

is a orthonormal set (this operation is called vector normalization). Proof
Let's check that the $\mathbf{u}_{i}^{\prime}$ vectors are orthogonal:

$$
\mathbf{u}_{i}^{\prime} \cdot \mathbf{u}_{j}^{\prime}=\frac{\mathbf{u}_{i}}{\left\|\mathbf{u}_{i}\right\|} \cdot \frac{\mathbf{u}_{j}}{\left\|\mathbf{u}_{j}\right\|}=\frac{1}{\left\|\mathbf{u}_{i}\right\|\left\|\mathbf{u}_{j}\right\|} \mathbf{u}_{i} \cdot \mathbf{u}_{j}
$$

But this product is obviusly 0 because the $\mathbf{u}_{i}$ vectors are orthogonal. Let's check now that the $\mathbf{u}_{i}^{\prime}$ vectors have unit length:

$$
\left\|\mathbf{u}_{i}^{\prime}\right\|=\left\|\frac{\mathbf{u}_{i}}{\left\|\mathbf{u}_{i}\right\|}\right\|=\frac{\left\|\mathbf{u}_{i}\right\|}{\left\|u_{i}\right\|}=1
$$

## Orthonormal matrix

## Theorem 2.5

Let $U \in \mathcal{M}_{m \times n}$ be a square matrix. The columns of $U$ form an orthonormal set iff

$$
U^{T} U=I_{n}
$$

It is said that $U$ is an orthonormal matrix.
Proof
Let's consider the columns of $U$

$$
U=\left(\begin{array}{llll}
\mathbf{u}_{1} & \mathbf{u}_{2} & \ldots & \mathbf{u}_{n}
\end{array}\right)
$$

Let's calculate now $U^{T} U$

$$
U^{T} U=\left(\begin{array}{c}
\mathbf{u}_{1}^{T} \\
\mathbf{u}_{2}^{T} \\
\ldots \\
\mathbf{u}_{n}^{T}
\end{array}\right)\left(\begin{array}{llll}
\mathbf{u}_{1} & \mathbf{u}_{2} & \ldots & \mathbf{u}_{n}
\end{array}\right)=\left(\begin{array}{cccc}
\mathbf{u}_{1}^{T} \mathbf{u}_{1} & \mathbf{u}_{1}^{T} \mathbf{u}_{2} & \ldots & \mathbf{u}_{1}^{T} \mathbf{u}_{n} \\
\mathbf{u}_{2}^{T} \mathbf{u}_{1} & \mathbf{u}_{2}^{T} \mathbf{u}_{2} & \ldots & \mathbf{u}_{2}^{T} \mathbf{u}_{n} \\
\ldots & \ldots & \ldots & \ldots \\
\mathbf{u}_{n}^{T} \mathbf{u}_{1} & \mathbf{u}_{n}^{T} \mathbf{u}_{2} & \ldots & \mathbf{u}_{n}^{T} \mathbf{u}_{n}
\end{array}\right)
$$

The condition $U^{T} U=I_{n}$ simply states $\left\{\begin{array}{ll}\mathbf{u}_{i}^{T} \mathbf{u}_{j}=0 & i \neq j \\ \mathbf{u}_{i}^{T} \mathbf{u}_{j}=1 & i=j\end{array}\right.$, which is the definition of an orthonormal set.

## Orthonormal matrix

## Theorem 2.6

Let $U \in \mathcal{M}_{n \times n}$ be an orthonormal matrix and $\forall \mathbf{x}, \mathbf{y} \in \mathbb{R}^{n}$, then

- $\|U \mathbf{x}\|=\|\mathbf{x}\|$
(2) $(U \mathbf{x}) \cdot(U \mathbf{y})=\mathbf{x} \cdot \mathbf{y}$
(3) $(U \mathbf{x}) \cdot(U \mathbf{y})=0 \Leftrightarrow \mathbf{x} \cdot \mathbf{y}=0$


## Example

Let $U=\left(\begin{array}{cc}\frac{1}{\sqrt{2}} & \frac{2}{3} \\ \frac{1}{\sqrt{2}} & -\frac{2}{3} \\ 0 & \frac{1}{3}\end{array}\right)$ and $\mathbf{x}=\binom{\sqrt{2}}{3}$.
$U$ is an orthonormal matrix because

$$
U^{T} U=\left(\begin{array}{ccc}
\frac{1}{\sqrt{2}} & \frac{1}{\sqrt{2}} & 0 \\
\frac{2}{3} & & \\
& -\frac{2}{3} & \frac{1}{3}
\end{array}\right)\left(\begin{array}{cc}
\frac{1}{\sqrt{2}} & \frac{2}{3} \\
\frac{1}{\sqrt{2}} & -\frac{2}{3} \\
0 & \frac{1}{3}
\end{array}\right)=\left(\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right)
$$

## Orthonormal matrix

## Example (continued)

Let's calculate now Ux

$$
U x=\left(\begin{array}{cc}
\frac{1}{\sqrt{2}} & \frac{2}{3} \\
\frac{1}{\sqrt{2}} & -\frac{2}{3} \\
0 & \frac{1}{3}
\end{array}\right)\binom{\sqrt{2}}{3}=\left(\begin{array}{c}
3 \\
-1 \\
1
\end{array}\right)
$$

Let's check now that $\|U \mathbf{x}\|=\|\mathbf{x}\|$

$$
\begin{gathered}
\|U \mathbf{x}\|=\|(3,-1,1)\|=\sqrt{3^{2}+(-1)^{2}+1^{2}}=\sqrt{11} \\
\|\mathbf{x}\|=\|(\sqrt{2}, 3)\|=\sqrt{(\sqrt{2})^{2}+3^{2}}=\sqrt{11}
\end{gathered}
$$

## Theorem 2.7

Let $U$ be an orthonormal and square matrix. Then,
(1) $U^{-1}=U^{T}$
(2) $U^{\top}$ is also an orthonormal matrix (i.e., the rows of $U$ also form an orthonormal set of vectors).

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 6, Section 2:

- 6.2.1
- 6.2.10
- 6.2.15
- 6.2 .25
- 6.2.26
- 6.2.29
- 6.2.35 (computer)


## Outline

(7) Orthogonality and least squares

- Inner product, length and orthogonality (a)
- Orthogonal sets, bases and matrices (a)
- Orthogonal projections (b)
- Gram-Schmidt orthogonalization (b)
- Least squares (c)
- Least-squares linear regression (c)
- Inner product spaces (d)
- Applications of inner product spaces (d)


## Orthogonal projections

## Definition 3.1 (Orthogonal projection)

The orthogonal projection of a point $\mathbf{y}$ onto a vector subspace $W$ is a point $\hat{\mathbf{y}} \in W$ such that

$$
\begin{gathered}
\mathbf{z}=\mathbf{y}-\hat{\mathbf{y}} \\
\mathbf{z} \perp W
\end{gathered}
$$



## Orthogonal projections

## Example

Let $\left\{\mathbf{u}_{1}, \mathbf{u}_{2}, \ldots, \mathbf{u}_{5}\right\}$ be an orthogonal basis of $\mathbb{R}^{5}$. Consider the subspace $W=\operatorname{Span}\left\{\mathbf{u}_{1}, \mathbf{u}_{2}\right\}$. Given any vector $\mathbf{y} \in \mathbb{R}^{5}$, we can decompose it as the sum of a vector in $W$ and a vector perpendicular to $W$

$$
\mathbf{y}=\hat{\mathbf{y}}+\mathbf{z}
$$

## Solution

If $\left\{\mathbf{u}_{1}, \mathbf{u}_{2}, \ldots, \mathbf{u}_{5}\right\}$ is a basis of $\mathbb{R}^{5}$, then any vector $\mathbf{y} \in \mathbb{R}^{5}$ can be written as

$$
\mathbf{y}=c_{1} \mathbf{u}_{1}+c_{2} \mathbf{u}_{2}+\ldots+c_{5} \mathbf{u}_{5}
$$

We may decompose this sum as

$$
\begin{gathered}
\hat{\mathbf{y}}=c_{1} \mathbf{u}_{1}+c_{2} \mathbf{u}_{2} \\
\mathbf{z}=c_{3} \mathbf{u}_{3}+c_{4} \mathbf{u}_{4}+c_{5} \mathbf{u}_{5}
\end{gathered}
$$

## Orthogonal projections

## Example (continued)

It is obvious that $\hat{\mathbf{y}} \in W$. Now we need to show that $\mathbf{z} \in W^{\perp}$. For doing so, we will show that

$$
\begin{aligned}
& \mathbf{z} \cdot \mathbf{u}_{1}=0 \\
& \mathbf{z} \cdot \mathbf{u}_{2}=0
\end{aligned}
$$

To show the first equation we note that

$$
\begin{aligned}
\mathbf{z} \cdot \mathbf{u}_{1} & =\left(c_{3} \mathbf{u}_{3}+c_{4} \mathbf{u}_{4}+c_{5} \mathbf{u}_{5}\right) \cdot \mathbf{u}_{1} \\
& =c_{3}\left(\mathbf{u}_{3} \cdot \mathbf{u}_{1}\right)+c_{4}\left(\mathbf{u}_{4} \cdot \mathbf{u}_{1}\right)+c_{5}\left(\mathbf{u}_{5} \cdot \mathbf{u}_{1}\right) \\
& =c_{3} \cdot 0+c_{4} \cdot 0+c_{5} \cdot 0 \\
& =0
\end{aligned}
$$

We would proceed analogously for $\mathbf{z} \cdot \mathbf{u}_{2}=0$.

## Orthogonal projections

Theorem 3.1 (Orthogonal Decomposition Theorem)
Let $W$ be a vector subspace of a vector space $V$. Then, any vector $\mathbf{y} \in V$ can be written uniquely as

$$
\mathbf{y}=\hat{\mathbf{y}}+\mathbf{z}
$$

with $\hat{\mathbf{y}} \in W$ and $\mathbf{z} \in W^{\perp}$. In fact, if $\left\{\mathbf{u}_{1}, \mathbf{u}_{2}, \ldots, \mathbf{u}_{p}\right\}$ is an orthogonal basis of $W$, then

$$
\hat{\mathbf{y}}=\frac{\mathbf{y} \cdot \mathbf{u}_{1}}{\left\|\mathbf{u}_{1}\right\|^{2}} \mathbf{u}_{1}+\frac{\mathbf{y} \cdot \mathbf{u}_{2}}{\left\|\mathbf{u}_{2}\right\|^{2}} \mathbf{u}_{2}+\ldots+\frac{\mathbf{y} \cdot \mathbf{u}_{p}}{\left\|\mathbf{u}_{p}\right\|^{2}} \mathbf{u}_{p}
$$



## Orthogonal projections

## Proof

$\hat{\mathbf{y}}$ is obviously in $W$ since it has been written as a linear combination of vectors in a basis of $W . \mathbf{z}$ is perpendicular to $W$ because

$$
\begin{aligned}
\mathbf{z} \cdot \mathbf{u}_{1}= & \left(\mathbf{y}-\left(\frac{\mathbf{y} \cdot \mathbf{u}_{1}}{\left\|\mathbf{u}_{1}\right\|^{2}} \mathbf{u}_{1}+\frac{\mathbf{y} \cdot \mathbf{u}_{2}}{\left\|\mathbf{u}_{2}\right\|^{2}} \mathbf{u}_{2}+\ldots+\frac{\mathbf{y} \cdot \mathbf{u}_{p}}{\left\|\mathbf{u}_{p}\right\|^{2}} \mathbf{u}_{p}\right)\right) \cdot \mathbf{u}_{1} \\
& =\mathbf{y} \cdot \mathbf{u}_{1}-\frac{\mathbf{y} \cdot \mathbf{u}_{1}}{\left\|\mathbf{u}_{1}\right\|^{2}}\left(\mathbf{u}_{1} \cdot \mathbf{u}_{1}\right)-\frac{\mathbf{y} \cdot \mathbf{u}_{2}}{\left\|\mathbf{u}_{2}\right\|^{2}}\left(\mathbf{u}_{2} \cdot \mathbf{u}_{1}\right)-\ldots-\frac{\mathbf{y} \cdot \mathbf{u}_{p}}{\left\|\mathbf{u}_{p}\right\|^{2}}\left(\mathbf{u}_{p} \cdot \mathbf{u}_{1}\right) \\
& {\left[\left\{\mathbf{u}_{i}\right\} \text { is an orthogonal set }\right] } \\
= & \mathbf{y} \cdot \mathbf{u}_{1}-\frac{\mathbf{y} \cdot \mathbf{u}_{1}}{\left\|\mathbf{u}_{1}\right\|^{2}}\left(\mathbf{u}_{1} \cdot \mathbf{u}_{1}\right) \\
= & \mathbf{y} \cdot \mathbf{u}_{1}-\frac{\mathbf{y} \cdot \mathbf{u}_{1}}{\left\|\mathbf{u}_{1}\right\|^{2}}\left\|\mathbf{u}_{1}\right\|^{2} \\
= & \mathbf{y} \cdot \mathbf{u}_{1}-\mathbf{y} \cdot \mathbf{u}_{1} \\
= & 0
\end{aligned}
$$

We could proceed analogously for all elements in the basis of $W$.

## Orthogonal projections

We need to show now that the decomposition is unique. Let us assume that it is not unique. Consequently, there exist different vectors such that

$$
\begin{gathered}
\mathbf{y}=\hat{\mathbf{y}}+\mathbf{z} \\
\mathbf{y}=\hat{\mathbf{y}}^{\prime}+\mathbf{z}^{\prime}
\end{gathered}
$$

We subtract both equations

$$
\mathbf{0}=\left(\hat{\mathbf{y}}-\hat{\mathbf{y}}^{\prime}\right)+\left(\mathbf{z}-\mathbf{z}^{\prime}\right) \Rightarrow \hat{\mathbf{y}}-\hat{\mathbf{y}}^{\prime}=\mathbf{z}^{\prime}-\mathbf{z}
$$

Let $\mathbf{v}=\hat{\mathbf{y}}-\hat{\mathbf{y}}^{\prime}$. It is obvious that $\mathbf{v} \in W$ because it is written as a linear combination of vectors in $W$. On the other side, $\mathbf{v}=\mathbf{z}^{\prime}-\mathbf{z}$, i.e., it is a linear combination of vectors in $W^{\perp}$, so $\mathbf{v} \in W^{\perp}$. The only vector that belongs to $W$ and $W^{\perp}$ at the same time is

$$
\mathbf{v}=\mathbf{0} \Rightarrow\left\{\begin{array}{l}
\hat{\mathbf{y}}=\hat{\mathbf{y}}^{\prime} \\
\mathbf{z}=\mathbf{z}^{\prime}
\end{array}\right.
$$

and consequently, the orthogonal decomposition is unique. Additionally, the uniqueness of the decomposition depends only on $W$ and not on the particular basis chosen for $W$.

## Orthogonal projections

## Example

Let $\mathbf{u}_{1}=(2,5,-1)$ and $\mathbf{u}_{2}=(-2,1,1)$. Let $W$ be the subspace spanned by $\mathbf{u}_{1}$ and $\mathbf{u}_{2}$. Let $\mathbf{y}=(1,2,3) \in \mathbb{R}^{3}$. The orthogonal projection of $\mathbf{y}$ onto $W$ is

$$
\begin{aligned}
\hat{\mathbf{y}} & =\frac{\mathbf{y} \cdot \mathbf{u}_{1}}{\left\|\mathbf{u}_{1}\right\|^{2}} \mathbf{u}_{1}+\frac{\mathbf{y} \cdot \mathbf{u}_{2}}{\left\|\mathbf{u}_{2}\right\|^{2}} \mathbf{u}_{2} \\
& =\frac{1 \cdot 2+2 \cdot 5+3 \cdot(-1)}{2^{2}+5^{2}+(-1)^{2}}\left(\begin{array}{c}
2 \\
5 \\
-1
\end{array}\right)+\frac{1 \cdot(-2)+2 \cdot 1+3 \cdot 1}{(-2)^{2}+1^{2}+1^{2}}\left(\begin{array}{c}
-2 \\
1 \\
1
\end{array}\right) \\
& =\frac{9}{30}\left(\begin{array}{c}
2 \\
5 \\
-1
\end{array}\right)+\frac{15}{30}\left(\begin{array}{c}
-2 \\
1 \\
1
\end{array}\right)=\left(\begin{array}{c}
-\frac{2}{5} \\
2 \\
\frac{1}{5}
\end{array}\right) \\
\mathbf{z} & =\mathbf{y}-\hat{\mathbf{y}}=\left(\begin{array}{l}
1 \\
2 \\
3
\end{array}\right)-\left(\begin{array}{c}
-\frac{2}{5} \\
2 \\
\frac{1}{5}
\end{array}\right)=\left(\begin{array}{c}
\frac{7}{5} \\
0 \\
\frac{14}{5}
\end{array}\right)
\end{aligned}
$$

## Orthogonal projections

## Geometrical interpretation

$\hat{\mathbf{y}}$ can be understood as the sum of the orthogonal projection of $\mathbf{y}$ onto each one of the elements of the basis of $W$.


Theorem 3.2
If $\mathbf{y}$ belongs to $W$, then the orthogonal projection of $\mathbf{y}$ onto $W$ is itself:

$$
\hat{\mathbf{y}}=\mathbf{y}
$$

## Properties of orthogonal projections

## Theorem 3.3 (Best approximation theorem)

The orthogonal projection of $\mathbf{y}$ onto $W$ is the point in $W$ with minimum distance to $\mathbf{y}$, i.e.,

$$
\|\mathbf{y}-\hat{\mathbf{y}}\| \leq\|\mathbf{y}-\mathbf{v}\|
$$

for all $\mathbf{v} \in W, \mathbf{v} \neq \hat{\mathbf{y}}$.
Proof
We know that $\mathbf{y}-\hat{\mathbf{y}}$ is orthogonal to $W$. For any vector $\mathbf{v} \in W, \mathbf{v} \neq \hat{\mathbf{y}}$, we have that $\hat{\mathbf{y}}-\mathbf{v}$ is in $W$. Now consider the orthogonal decomposition of the vector $\mathbf{y}-\mathbf{v}$

$$
\mathbf{y}-\mathbf{v}=(\mathbf{y}-\hat{\mathbf{y}})+(\hat{\mathbf{y}}-\mathbf{v})
$$



FIGURE 4 The orthogonal projection of $\mathbf{y}$ onto $W$ is the closest point in $W$ to $\mathbf{y}$.

## Properties of orthogonal projections

Due to the orthogonal decomposition theorem (Theorem 3.1), this decomposition is unique and due to the Pythagorean theorem (Theorem 1.5) we have

$$
\|\mathbf{y}-\mathbf{v}\|^{2}=\|\mathbf{y}-\hat{\mathbf{y}}\|^{2}+\|\hat{\mathbf{y}}-\mathbf{v}\|^{2}
$$

Since $\mathbf{v} \neq \hat{\mathbf{y}}$ we have $\|\hat{\mathbf{y}}-\mathbf{v}\|^{2}>0$ and consequently

$$
\|\mathbf{y}-\mathbf{v}\|^{2}>\|\mathbf{y}-\hat{\mathbf{y}}\|^{2}
$$

## Properties of orthogonal projections

## Theorem 3.4

If $\left\{\mathbf{u}_{1}, \mathbf{u}_{2}, \ldots, \mathbf{u}_{p}\right\}$ is an orthonormal basis of $W$, then the orthogonal projection of y onto $W$ is

$$
\hat{\mathbf{y}}=\left\langle\mathbf{y}, \mathbf{u}_{1}\right\rangle \mathbf{u}_{1}+\left\langle\mathbf{y}, \mathbf{u}_{2}\right\rangle \mathbf{u}_{2}+\ldots+\left\langle\mathbf{y}, \mathbf{u}_{p}\right\rangle \mathbf{u}_{p}
$$

If we construct the orthonormal matrix $U=\left(\begin{array}{llll}\mathbf{u}_{1} & \mathbf{u}_{2} & \ldots & \mathbf{u}_{p}\end{array}\right)$, then

$$
\hat{\mathbf{y}}=U U^{T} \mathbf{y}
$$

## Proof

By Theorem 3.1 we know that for all orthogonal bases it is verified

$$
\hat{\mathbf{y}}=\frac{\mathrm{y} \cdot \mathbf{u}_{1}}{\| \mathbf{u}_{1} \mathbf{u}^{2}} \mathbf{u}_{1}+\frac{\mathrm{y} \cdot \mathbf{u}_{2}}{\left\|\mathbf{u}_{2}\right\|^{2}} \mathbf{u}_{2}+\ldots+\frac{\mathbf{y} \cdot \mathbf{u}_{p}}{\left\|\mathbf{u}_{p}\right\|^{2}} \mathbf{u}_{p}
$$

Since the basis is in this case orthonormal, then $\|\mathbf{u}\|=1$ and consequently

$$
\hat{\mathbf{y}}=\left\langle\mathbf{y}, \mathbf{u}_{1}\right\rangle \mathbf{u}_{1}+\left\langle\mathbf{y}, \mathbf{u}_{2}\right\rangle \mathbf{u}_{2}+\ldots+\left\langle\mathbf{y}, \mathbf{u}_{p}\right\rangle \mathbf{u}_{p}
$$

## Properties of orthogonal projections

On the other side we have

$$
U^{T} \mathbf{y}=\left(\begin{array}{c}
\mathbf{u}_{1}^{T} \\
\mathbf{u}_{2}^{T} \\
\ldots \\
\mathbf{u}_{p}^{T}
\end{array}\right) \mathbf{y}=\left(\begin{array}{c}
\mathbf{u}_{1}^{T} \mathbf{y} \\
\mathbf{u}_{2}^{T} \mathbf{y} \\
\ldots \\
\mathbf{u}_{\rho}^{T} \mathbf{y}
\end{array}\right)=\left(\begin{array}{c}
\left\langle\mathbf{u}_{1}, \mathbf{y}\right\rangle \\
\left\langle\mathbf{u}_{2}, \mathbf{y}\right\rangle \\
\ldots \\
\left\langle\mathbf{u}_{\rho}, \mathbf{y}\right\rangle
\end{array}\right)
$$

Then,

$$
U U^{T} \mathbf{y}=\left(\begin{array}{llll}
\mathbf{u}_{1} & \mathbf{u}_{2} & \ldots & \mathbf{u}_{p}
\end{array}\right)\left(\begin{array}{c}
\left\langle\mathbf{u}_{1}, \mathbf{y}\right\rangle \\
\left\langle\mathbf{u}_{2}, \mathbf{y}\right\rangle \\
\ldots \\
\left\langle\mathbf{u}_{p}, \mathbf{y}\right\rangle
\end{array}\right)=\left\langle\mathbf{y}, \mathbf{u}_{1}\right\rangle \mathbf{u}_{1}+\left\langle\mathbf{y}, \mathbf{u}_{2}\right\rangle \mathbf{u}_{2}+\ldots+\left\langle\mathbf{y}, \mathbf{u}_{p}\right\rangle \mathbf{u}_{p}
$$

(q.e.d.)

## Properties of orthogonal projections

## Corollary

Let $U=\left(\begin{array}{llll}\mathbf{u}_{1} & \mathbf{u}_{2} & \ldots & \mathbf{u}_{p}\end{array}\right)$ be a $n \times p$ matrix with orthonormal columns and $W=\operatorname{Col}\{U\}$ its column space. Then,

$$
\begin{array}{lll}
\forall \mathbf{x} \in \mathbb{R}^{p} & U^{T} U \mathbf{x}=\mathbf{x} & \text { No effect } \\
\forall \mathbf{y} \in \mathbb{R}^{n} & U U^{T} \mathbf{y}=\hat{\mathbf{y}} & \text { Orthogonal projection of } \mathbf{y} \text { onto } W
\end{array}
$$

If $U$ is a $n \times n$, then $W=\mathbb{R}^{n}$ and the projection has no effect

$$
\forall \mathbf{y} \in \mathbb{R}^{n} \quad U U^{T} \mathbf{y}=\hat{\mathbf{y}}=\mathbf{y} \quad \text { No effect }
$$

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 6, Section 3:

- 6.3.1
- 6.3.7
- 6.3.15
- 6.3.23
- 6.3.24
- 6.3.25 (computer)


## Outline

(7) Orthogonality and least squares

- Inner product, length and orthogonality (a)
- Orthogonal sets, bases and matrices (a)
- Orthogonal projections (b)
- Gram-Schmidt orthogonalization (b)
- Least squares (c)
- Least-squares linear regression (c)
- Inner product spaces (d)
- Applications of inner product spaces (d)


## Gram-Schmidt orthogonalization

Gram-Schmidt orthogonalization is a procedure aimed at producing an orthogonal basis of any subspace $W$.

## Example

Let $W=\operatorname{Span}\left\{\mathbf{x}_{1}, \mathbf{x}_{2}\right\}$ with $\mathbf{x}_{1}=(3,6,0)$ and $\mathbf{x}_{2}=(1,2,2)$. Let's look for an orthogonal basis of $W$.

## Solution

We may keep the first vector for the basis

$$
\mathbf{v}_{1}=\mathbf{x}_{1}=(3,6,0)
$$

For the second vector in the basis, we need to keep the component of $\mathbf{x}_{2}$ that is orthogonal to $\mathbf{x}_{1}$. For doing so we calculate the projection of $\mathbf{x}_{2}$ onto $\mathbf{x}_{1}(\mathbf{p})$, and we decompose $\mathbf{x}_{2}$ as

$$
\mathbf{x}_{2}=\mathbf{p}+\left(\mathbf{x}_{2}-\mathbf{p}\right)=(1,2,0)+(0,0,2)
$$

We, then, keep the orthogonal part of $\mathbf{x}_{2}$

$$
\mathbf{v}_{2}=\mathbf{x}_{2}-\mathbf{p}=(0,0,2)
$$

## Gram-Schmidt orthogonalization

## Example (continued)

The set $\left\{\mathbf{v}_{1}, \mathbf{v}_{2}\right\}$ is an orthogonal basis of $W$.


FIGURE 1
Construction of an orthogonal basis $\left\{\mathbf{v}_{1}, \mathbf{v}_{2}\right\}$.

## Gram-Schmidt orthogonalization

## Example

Let $W=\operatorname{Span}\left\{\mathbf{x}_{1}, \mathbf{x}_{2}, \mathbf{x}_{3}\right\}$ with $\mathbf{x}_{1}=(1,1,1,1), \mathbf{x}_{2}=(0,1,1,1)$ and $x_{3}=(0,0,1,1)$. Let's look for an orthogonal basis of $W$.

## Solution

We may keep the first vector for the basis. Then we construct a subspace ( $W_{1}$ ) with a single element in its basis

$$
\mathbf{v}_{1}=\mathbf{x}_{1}=(1,1,1,1) \quad W_{1}=\operatorname{Span}\left\{\mathbf{v}_{1}\right\}
$$

For the second vector in the basis, we need to keep the component of $\mathbf{x}_{2}$ that is orthogonal to $W_{1}$. With the already computed basis vectors, we construct a new subspace ( $W_{2}$ ) with two elements in its basis

$$
\mathbf{v}_{2}=\mathbf{x}_{2}-\operatorname{Proj}_{W_{1}}\left(\mathbf{x}_{2}\right)=\left(-\frac{3}{4}, \frac{1}{4}, \frac{1}{4}, \frac{1}{4}\right) \quad W_{2}=\operatorname{Span}\left\{\mathbf{v}_{1}, \mathbf{v}_{2}\right\}
$$

For the third vector in the basis, we repeat the same procedure

$$
\mathbf{v}_{3}=\mathbf{x}_{3}-\operatorname{Proj}_{W_{2}}\left(\mathbf{x}_{3}\right)=\left(0,-\frac{2}{3}, \frac{1}{3}, \frac{1}{3}\right) \quad W_{3}=\operatorname{Span}\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \mathbf{v}_{3}\right\}
$$

## Gram-Schmidt orthogonalization

## Theorem 4.1 (Gram-Schmidt orthogonalization)

Given a basis $\left\{\mathbf{x}_{1}, \mathbf{x}_{2}, \ldots, \mathbf{x}_{p}\right\}$ for a vector subspace W. Define

$$
\begin{array}{cl}
\mathbf{v}_{1}=\mathbf{x}_{1} & W_{1}=\operatorname{Span}\left\{\mathbf{v}_{1}\right\} \\
\mathbf{v}_{2}=\mathbf{x}_{2}-\operatorname{Proj}_{W_{1}}\left(\mathbf{x}_{2}\right) & W_{2}=\operatorname{Span}\left\{\mathbf{v}_{1}, \mathbf{v}_{2}\right\} \\
\ldots & \\
\mathbf{v}_{p}=\mathbf{x}_{\rho}-\operatorname{Proj}_{W_{p-1}}\left(\mathbf{x}_{\rho}\right) & W_{p}=\operatorname{Span}\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{p}\right\}=W
\end{array}
$$

Then $\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{p}\right\}$ is an orthogonal basis of $W$.

## Gram-Schmidt orthogonalization

## Proof

Consider $W_{k}=\operatorname{Span}\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{k}\right\}$ and let us assume that $\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{k}\right\}$ is a basis of $W_{k}$. Now we construct

$$
\mathbf{v}_{k+1}=\mathbf{x}_{k+1}-\operatorname{Proj}_{W_{k}}\left(\mathbf{x}_{k+1}\right) \quad W_{k+1}=\operatorname{Span}\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{k+1}\right\}
$$

By the orthogonal decomposition theorem (Theorem 3.1), we know that $\mathbf{v}_{k+1}$ is orthogonal to $W_{k}$. Because $\mathbf{x}_{k+1}$ is an element of a basis, we know that $\mathbf{x}_{k+1} \notin W_{k}$. Therefore, $\mathbf{v}_{k+1}$ is not null and $\mathbf{x}_{k+1} \in W_{k+1}$. Finally, the set $\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{k+1}\right\}$ is a set of orthogonal, non-null vectors in the $(k+1)$-dimensional space $W_{k+1}$. Consequently, by Theorem 9.4 in Chapter 5 , it must be a basis of $W_{k+1}$. This process can be iterated till $k=p$.

## Gram-Schmidt orthogonalization

## Orthonormal basis

Once we have an orthogonal basis, we simply have to normalize each vector to have an orthonormal basis.

## Example

Let $W=\operatorname{Span}\left\{\mathbf{x}_{1}, \mathbf{x}_{2}\right\}$ with $\mathbf{x}_{1}=(3,6,0)$ and $\mathbf{x}_{2}=(1,2,2)$. Let's look for an orthonormal basis of $W$.

## Solution

In Slide 52 we learned that an orthogonal basis was given by

$$
\begin{aligned}
& \mathbf{v}_{1}=(3,6,0) \\
& \mathbf{v}_{2}=(0,0,2)
\end{aligned}
$$

Now, we normalize these two vectors to obtain an orthonormal basis

$$
\begin{gathered}
\mathbf{v}_{1}^{\prime}=\frac{\mathbf{v}_{1}}{\left\|\mathbf{v}_{1}\right\|}=\frac{1}{\sqrt{45}}(3,6,0)=\left(\frac{1}{\sqrt{5}}, \frac{2}{\sqrt{5}}, 0\right) \\
\mathbf{v}_{2}^{\prime}=\frac{\mathbf{v}_{2}}{\left\|\mathbf{v}_{2}\right\|}=\frac{1}{2}(0,0,2)=(0,0,1)
\end{gathered}
$$

## QR factorization of matrices

If we apply the Gram-Schmidt factorization to the columns of a matrix, we have the following factorization scheme. This factorization is used in practice to find eigenvalues and eigenvectors as well as to solve linear equation systems.

## Theorem 4.2 (QR factorization)

Let $A \in \mathcal{M}_{m \times n}$ with linearly independent columns. Then, $A$ can be factored as

$$
A=Q R
$$

where $Q \in \mathcal{M}_{m \times n}$ is a matrix whose columns form an orthonormal basis of $\operatorname{Col}\{A\}$ and $R \in \mathcal{M}_{n \times n}$ is an upper triangular invertible matrix with positive entries on its diagonal.
Proof
Let's orthogonalize the columns of A following the Gram-Schmidt procedure and construct the orthonormal basis of $\operatorname{Col}\{A\}$. Let $\left\{\mathbf{u}_{1}, \mathbf{u}_{2}, \ldots, \mathbf{u}_{n}\right\}$ be such a basis. Let us construct the matrix

$$
Q=\left(\begin{array}{llll}
\mathbf{u}_{1} & \mathbf{u}_{2} & \ldots & \mathbf{u}_{n}
\end{array}\right)
$$

## QR factorization of matrices

Let us call $\mathbf{a}_{i}(i=1,2, \ldots, n)$ to the columns of $A$. By the Gram-Schmidt orthogonalization, we know that for any $k$ between 1 and $n$ we have

$$
\operatorname{Span}\left\{\mathbf{a}_{1}, \mathbf{a}_{2}, \ldots, \mathbf{a}_{k}\right\}=\operatorname{Span}\left\{\mathbf{u}_{1}, \mathbf{u}_{2}, \ldots, \mathbf{u}_{k}\right\}
$$

Consequently, we can express each column of $A$ in the orthonormal basis:

$$
\mathbf{a}_{k}=r_{1 k} \mathbf{u}_{1}+r_{2 k} \mathbf{u}_{2}+\ldots+r_{k k} \mathbf{u}_{k}+0 \cdot \mathbf{u}_{k+1}+\ldots+0 \cdot \mathbf{u}_{n}
$$

If $r_{k k}$ is negative, we can multiply both $r_{k k}$ and $\mathbf{u}_{k}$ by -1 . We now collect all these coefficients in a vector $\mathbf{r}_{k}=\left(r_{1 k}, r_{2 k}, \ldots, r_{k k}, 0,0, \ldots, 0\right)$ to have

$$
\mathbf{a}_{k}=Q \mathbf{r}_{k}
$$

By gathering all these vectors in a matrix, we have the triangular matrix $R$

$$
R=\left(\begin{array}{llll}
\mathbf{r}_{1} & \mathbf{r}_{2} & \ldots & \mathbf{r}_{n}
\end{array}\right)
$$

$R$ is invertible because the columns of $A$ are linearly independent.

## QR factorization of matrices

## Example

Let's calculate the QR factorization of $A=\left(\begin{array}{lll}1 & 0 & 0 \\ 1 & 1 & 0 \\ 1 & 1 & 1 \\ 1 & 1 & 1\end{array}\right)$. From Slide 54 we know that the vectors

$$
\begin{gathered}
\mathbf{v}_{1}=(1,1,1,1) \\
\mathbf{v}_{2}=\left(-\frac{3}{4}, \frac{1}{4}, \frac{1}{4}, \frac{1}{4}\right) \\
\mathbf{v}_{3}=\left(0,-\frac{2}{3}, \frac{1}{3}, \frac{1}{3}\right)
\end{gathered}
$$

Is an orthogonal basis of the column space of $A$. We now normalize these vectors to obtain the orthonormal basis in $Q$

$$
Q=\left(\begin{array}{ccc}
\frac{1}{2} & -\frac{3}{\sqrt{12}} & 0 \\
\frac{1}{2} & \frac{1}{\sqrt{12}} & -\frac{2}{\sqrt{6}} \\
\frac{1}{2} & \frac{1}{\sqrt{12}} & \frac{1}{\sqrt{6}} \\
\frac{1}{2} & \frac{1}{\sqrt{12}} & \frac{1}{\sqrt{6}}
\end{array}\right)
$$

## QR factorization of matrices

## Example (continued)

To find $R$ we multiply on both sides of the factorization by $Q^{T}$

$$
\begin{aligned}
& A=Q R \Rightarrow Q^{T} A=Q^{T} Q R=R \\
R=Q^{T} A= & \left(\begin{array}{cccc}
\frac{1}{2} & \frac{1}{2} & \frac{1}{2} & \frac{1}{2} \\
-\frac{3}{\sqrt{12}} & \frac{1}{\sqrt{12}} & \frac{1}{\sqrt{12}} & \frac{1}{\sqrt{12}} \\
0 & -\frac{2}{\sqrt{6}} & \frac{1}{\sqrt{6}} & \frac{1}{\sqrt{6}}
\end{array}\right)\left(\begin{array}{ccc}
1 & 0 & 0 \\
1 & 1 & 0 \\
1 & 1 & 1 \\
1 & 1 & 1
\end{array}\right) \\
= & \left(\begin{array}{ccc}
2 & \frac{3}{2} & 1 \\
0 & \frac{3}{\sqrt{12}} & \frac{2}{\sqrt{12}} \\
0 & 0 & \frac{1}{\sqrt{6}}
\end{array}\right)
\end{aligned}
$$

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 6, Section 4:

- 6.4.7
- 6.4.13
- 6.4.19
- 6.4.22
- 6.4.24


## Outline

(7) Orthogonality and least squares

- Inner product, length and orthogonality (a)
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## Least squares

Let's assume we want to solve the equation system $A \mathbf{x}=\mathbf{b}$, but, due to noise, there is no solution. We may still look for a solution such that $A \mathbf{x} \approx \mathbf{b}$. In fact the goal will be to minimize $d(A \mathbf{x}, \mathbf{b})$.

## Definition 5.1 (Least squares solution)

Let $A$ be a $m \times n$ matrix and $\mathbf{b} \in \mathbb{R}^{m} . \mathbf{x} \in \mathbb{R}^{n}$ is a least squares solution of the equation system $A \mathbf{x}=\mathbf{b}$ iff

$$
\forall \mathbf{x} \in \mathbb{R}^{n} \quad\|\mathbf{b}-A \hat{\mathbf{x}}\| \leq\|\mathbf{b}-A \mathbf{x}\|
$$



FIGURE 1 The vector $\mathbf{b}$ is closer to $A \hat{\mathbf{x}}$ than to $A \mathbf{x}$ for other $\mathbf{x}$.

## Least squares

## Solution of the general least squares problem

Applying the Best Approximation Theorem (Theorem 3.3), we may project b onto the column space of $A$

$$
\hat{\mathbf{b}}=\operatorname{Proj}_{\operatorname{Col}\{A\}}\{\mathbf{b}\}
$$

Then, we solve the equation system

$$
A \mathbf{x}=\hat{\mathbf{b}}
$$

that has at least one solution.


FIGURE 2 The least-squares solution $\hat{\mathbf{x}}$ is in $\mathbb{R}^{n}$.

## Least squares

## Theorem 5.1

The set of least-squares solutions of $A \mathbf{x}=\mathbf{b}$ is the same as the set of solutions of the normal equations

$$
A^{T} A \mathbf{x}=A^{T} \mathbf{b}
$$

Proof: least-squares solutions $\subset$ normal equations solutions
Let us assume that $\hat{\mathbf{x}}$ is a least-squares solution. Then, $\mathbf{b}-A \hat{\mathbf{x}}$ is orthogonal to $\operatorname{Col}\{A\}$, and in particular, to each one of the columns of $A\left(\mathbf{a}_{i}, i=1,2, \ldots, n\right)$ :

$$
\begin{gathered}
\mathbf{a}_{i} \cdot(\mathbf{b}-A \hat{\mathbf{x}})=0 \quad \forall i \in\{1,2, \ldots, n\} \Rightarrow \\
\mathbf{a}_{i}^{T}(\mathbf{b}-A \hat{\mathbf{x}})=0 \quad \forall i \in\{1,2, \ldots, n\} \Rightarrow \\
A^{T}(\mathbf{b}-A \hat{\mathbf{x}})=\mathbf{0} \Rightarrow \\
A^{T} \mathbf{b}=A^{T} A \hat{\mathbf{x}}
\end{gathered}
$$

That is, every least-squares solution is also a solution of the normal equations.

## Least squares

Proof: least-squares solutions $\supset$ normal equations solutions Let us assume that $\hat{\mathbf{x}}$ is solution of the normal equations. Then,

$$
\begin{gathered}
A^{T} \mathbf{b}=A^{T} A \hat{\mathbf{x}} \Rightarrow \\
A^{T}(\mathbf{b}-A \hat{\mathbf{x}})=0 \Rightarrow \\
\mathbf{a}_{i}^{T}(\mathbf{b}-A \hat{\mathbf{x}})=0 \quad \forall i \in\{1,2, \ldots, n\}
\end{gathered}
$$

That is, $\mathbf{b}-A \hat{\mathbf{x}}$ is orthogonal to the columns of $A$ and, consequently, to $\operatorname{Col}\{A\}$. Hence, the equation

$$
\mathbf{b}=A \hat{\mathbf{x}}+(\mathbf{b}-A \hat{\mathbf{x}})
$$

is the orthogonal decomposition of $\mathbf{b}$ as a vector in $\operatorname{Col}\{A\}$ and a vector orthogonal to $\operatorname{Col}\{A\}$. By the uniqueness of the orthogonal decomposition, $A \hat{\mathrm{x}}$ must be the orthogonal projection of $\mathbf{b}$ onto $\operatorname{Col}\{A\}$ so that

$$
A \hat{\mathbf{x}}=\hat{\mathbf{b}}
$$

and, therefore, $\hat{\mathrm{x}}$ is a least-squares solution.

## Least squares

## Example

Find a least-squares solution to $A \mathbf{x}=\mathbf{b}$ with $A=\left(\begin{array}{ll}4 & 0 \\ 0 & 2 \\ 1 & 1\end{array}\right)$ and $\mathbf{b}=\left(\begin{array}{c}2 \\ 0 \\ 11\end{array}\right)$.

## Solution

Let's solve the normal equations $A^{T} A \hat{\mathbf{x}}=A^{T} \mathbf{b}$

$$
\begin{gathered}
A^{T} A=\left(\begin{array}{cc}
17 & 1 \\
1 & 5
\end{array}\right) \quad A^{T} \mathbf{b}=\binom{19}{11} \\
\left(\begin{array}{cc}
17 & 1 \\
1 & 5
\end{array}\right) \hat{\mathbf{x}}=\binom{19}{11} \Rightarrow \hat{\mathbf{x}}=\left(\begin{array}{cc}
17 & 1 \\
1 & 5
\end{array}\right)^{-1}\binom{19}{11}=\binom{1}{2}
\end{gathered}
$$

Let's check that $\hat{\mathbf{x}}$ is not a solution of the original equation system but a least-squares solution

$$
A \hat{\mathbf{x}}=\left(\begin{array}{ll}
4 & 0 \\
0 & 2 \\
1 & 1
\end{array}\right)\binom{1}{2}=\left(\begin{array}{l}
4 \\
4 \\
3
\end{array}\right)=\hat{\mathbf{b}} \neq \mathbf{b}=\left(\begin{array}{c}
2 \\
0 \\
11
\end{array}\right)
$$

## Least squares

## Definition 5.2 (Least-squares error)

The least-squares error is defined as

$$
\sigma_{\epsilon}^{2} \triangleq\|A \hat{\mathbf{x}}-\mathbf{b}\|^{2}=\|\hat{\mathbf{b}}-\mathbf{b}\|^{2}
$$

## Example (continued)

In this case:

$$
\sigma_{\epsilon}^{2}=\|(4,4,3)-(2,0,11)\|=\|(2,4,-8)\| \approx 9.165
$$

## Least squares

## Example

Unfortunately, the least-squares solution may not be unique as shown in the next example (arising in ANOVA). Find a least-squares solution to $A \mathbf{x}=\mathbf{b}$ with
$A=\left(\begin{array}{llll}1 & 1 & 0 & 0 \\ 1 & 1 & 0 & 0 \\ 1 & 0 & 1 & 0 \\ 1 & 0 & 1 & 0 \\ 1 & 0 & 0 & 1 \\ 1 & 0 & 0 & 1\end{array}\right)$ and $\mathbf{b}=\left(\begin{array}{c}-3 \\ -1 \\ 0 \\ 2 \\ 5 \\ 1\end{array}\right)$.

## Solution

$$
A^{T} A=\left(\begin{array}{llll}
6 & 2 & 2 & 2 \\
2 & 2 & 0 & 0 \\
2 & 0 & 2 & 0 \\
2 & 0 & 0 & 2
\end{array}\right) \quad A^{T} \mathbf{b}=\left(\begin{array}{c}
4 \\
-4 \\
2 \\
6
\end{array}\right)
$$

## Least squares

## Example (continued)

The augmented matrix is

$$
\left(\begin{array}{rrrr|r}
6 & 2 & 2 & 2 & 4 \\
2 & 2 & 0 & 0 & -4 \\
2 & 0 & 2 & 0 & 2 \\
2 & 0 & 0 & 2 & 6
\end{array}\right) \sim\left(\begin{array}{rrrr|r}
1 & 0 & 0 & 1 & 3 \\
0 & 1 & 0 & -1 & -5 \\
0 & 0 & 1 & -1 & -2 \\
0 & 0 & 0 & 0 & 0
\end{array}\right)
$$

Any point of the form

$$
\hat{\mathbf{x}}=\left(\begin{array}{c}
3 \\
-5 \\
-2 \\
0
\end{array}\right)+x_{4}\left(\begin{array}{c}
-1 \\
1 \\
1 \\
1
\end{array}\right) \quad \forall x_{4} \in \mathbb{R}
$$

is a least-squares solution of the problem.

## Least squares

## Theorem 5.2

The matrix $A^{T} A$ is invertible iff the columns of $A$ are linearly independent. In this case, the equation system $A \mathbf{x}=\mathbf{b}$ has a unique least-squares solution given by

$$
\hat{\mathbf{x}}=A^{+} \mathbf{b}
$$

where $A^{+}$is the Moore-Penrose pseudoinverse

$$
A^{+}=\left(A^{T} A\right)^{-1} A^{T}
$$

## Least squares and QR decomposition

Sometimes $A^{T} A$ is ill-conditioned, this means that small perturbations in $\mathbf{b}$ translate into large perturbations in $\hat{\mathbf{x}}$. The QR decomposition offers a numerically more stable way of finding the least-squares solution.

## Theorem 5.3

Let there be $A \in \mathcal{M}_{m \times n}$ with linearly independent columns. Consider its $Q R$ decomposition $(A=Q R)$. Then, for each $\mathbf{b} \in \mathbb{R}^{m}$ there is a unique least-squares solution of $A \mathbf{x}=\mathbf{b}$ given by

$$
\hat{\mathbf{x}}=R^{-1} Q^{\top} \mathbf{b}
$$

Proof
If we substitute $\hat{\mathbf{x}}=R^{-1} Q^{\top} \mathbf{b}$ into $A \mathbf{x}$ we have

$$
A \hat{\mathbf{x}}=A R^{-1} Q^{\top} \mathbf{b}=Q R R^{-1} Q^{\top} \mathbf{b}=Q Q^{\top} \mathbf{b} .
$$

But $Q$ is an orthonormal basis of $\operatorname{Col}\{A\}$ (Theorem 4.2 and Corollary in Slide 49) and consequently $Q Q^{\top} \mathbf{b}$ is the orthogonal projection of $\mathbf{b}$ onto $\operatorname{Col}\{A\}$, that is, $\hat{\mathbf{b}}$. So, $\hat{\mathbf{x}}=R^{-1} Q^{\top} \mathbf{b}$ is a least-squares solution of $A \mathbf{x}=\mathbf{b}$. Additionally, since the columns of A are linearly independent, by Theorem 5.2, this solution is unique.

## Least squares and $Q R$ decomposition

Remind that numerically it is easier to solve $R \hat{\mathbf{x}}=Q^{\top} \mathbf{b}$ than $\hat{\mathbf{x}}=R^{-1} Q^{\top} \mathbf{b}$
L
et $A=\left(\begin{array}{lll}1 & 3 & 5 \\ 1 & 1 & 0 \\ 1 & 1 & 2 \\ 1 & 3 & 3\end{array}\right)$ and $\mathbf{b}=\left(\begin{array}{c}3 \\ 5 \\ 7 \\ -3\end{array}\right)$. Its QR decomposition is

$$
\begin{gathered}
A=Q R=\left(\begin{array}{rrr}
\frac{1}{2} & \frac{1}{2} & \frac{1}{2} \\
\frac{1}{2} & -\frac{1}{2} & -\frac{1}{2} \\
\frac{1}{2} & -\frac{1}{2} & \frac{1}{2} \\
\frac{1}{2} & \frac{1}{2} & -\frac{1}{2}
\end{array}\right)\left(\begin{array}{lll}
2 & 4 & 5 \\
0 & 2 & 3 \\
0 & 0 & 2
\end{array}\right) \\
Q^{\top} \mathbf{b}=\left(\begin{array}{r}
6 \\
-6 \\
4
\end{array}\right) \Rightarrow\left(\begin{array}{lll}
2 & 4 & 5 \\
0 & 2 & 3 \\
0 & 0 & 2
\end{array}\right) \hat{\mathbf{x}}=\left(\begin{array}{r}
6 \\
-6 \\
4
\end{array}\right) \Rightarrow \hat{\mathbf{x}}=\left(\begin{array}{r}
10 \\
-6 \\
2
\end{array}\right)
\end{gathered}
$$

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 6, Section 5:

- 6.5.1
- 6.5.19
- 6.5.20
-6.5.21
- 6.5.24


## Outline

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## Least-squares linear regression

## Example

In many scientific and engineering problems, it is needed to explain some observations $\mathbf{y}$ as a linear function of an independent variable $\mathbf{x}$. For instance, we may try to explain the weight of a person as a linear function of its height

$$
\text { Weight }=\beta_{0}+\beta_{1} \text { Height }
$$


A. Schneider, G. Hommel, M. Blettner. Linear Regression Analysis. Dtsch Arztebl Int. 2010 November; 107(44): 776-782.

## Least-squares linear regression

## Example (continued)

For each observation we have an equation

| Height (m.) | Weight (kg.) | $57=\beta_{0}+1.70 \beta_{1}$ <br> 1.70$\| 57$ |
| :---: | :---: | :---: |
| 1.53 | 43 | $43=\beta_{0}+1.53 \beta_{1}$ |
| 1.90 | 94 | $94=\beta_{0}+1.90 \beta_{1}$ |
| $\ldots$ | $\ldots$ | $\ldots$ |
|  | $\left(\begin{array}{cc}1 & 1.70 \\ 1 & 1.53 \\ 1 & 1.90 \\ \ldots & \ldots\end{array}\right)\binom{\beta_{0}}{\beta_{1}}=\left(\begin{array}{c}57 \\ 43 \\ 94 \\ \ldots\end{array}\right)$ |  |

which is of the form

$$
X \beta=\mathbf{y}
$$

## Least-squares linear regression

## Least-squares regression

Each one of the observed data points $\left(x_{j}, y_{j}\right)$ gives an equation. All together provide an equation system

$$
X \beta=\mathbf{y}
$$

that is an overdetermined, linear equation system of the form $A \mathbf{x}=\mathbf{b}$. The matrix $X$ is called the system matrix and it is related to the independent (predictor) variables (the height in this case). The vector $\mathbf{y}$ is called the observation vector and collects the values of the dependent (predicted) variable (the weight in this case). The model

$$
y=\beta_{0}+\beta_{1} x+\epsilon
$$

is called the linear regression of $y$ on $x . \beta_{0}$ and $\beta_{1}$ are called the regression coefficients. The difference between the predicted value and the observed value for a particular observation $(\epsilon)$ is called the residual of that observation.

## Least-squares linear regression



FIGURE 1 Fitting a line to experimental data.

The residual of the $j$-th observation is defined as

$$
\epsilon_{j}=y_{j}-\left(\beta_{0}+\beta_{1} x_{j}\right)
$$

## Least-squares linear regression

The goal of least-squares regression is to minimize

$$
\sum_{j=1}^{n} \epsilon_{j}^{2}=\|\mathbf{y}-X \beta\|^{2}
$$

Let's analyze this term

$$
X \beta=\left(\begin{array}{cc}
1 & x_{1} \\
1 & x_{2} \\
\ldots & \ldots \\
1 & x_{n}
\end{array}\right)\binom{\beta_{0}}{\beta_{1}}=\left(\begin{array}{c}
\beta_{0}+\beta_{1} x_{1} \\
\beta_{0}+\beta_{2} x_{2} \\
\ldots \\
\beta_{0}+\beta_{n} x_{n}
\end{array}\right)=\left(\begin{array}{c}
\hat{y}_{1} \\
\hat{y}_{2} \\
\ldots \\
\hat{y}_{n}
\end{array}\right)
$$

Then

$$
\|\mathbf{y}-x \beta\|^{2}=\left\|\left(\begin{array}{c}
y_{1}-\hat{y}_{1} \\
y_{2}-\hat{y}_{2} \\
\cdots \\
y_{n}-\hat{y}_{n}
\end{array}\right)\right\|^{2}=\sum_{j=1}^{n}\left(y_{j}-\hat{y}_{j}\right)^{2}=\sum_{j=1}^{n} \epsilon_{j}^{2}
$$

## Least-squares linear regression

## Example

Suppose we have observed the following values of height and weight $(1.70,57)$,
$(1.53,43),(1.90,94)$. We construct the system matrix $X=\left(\begin{array}{ll}1 & 1.70 \\ 1 & 1.53 \\ 1 & 1.90\end{array}\right)$ and the observation vector $\mathbf{y}=\left(\begin{array}{l}57 \\ 43 \\ 94\end{array}\right)$. Now we look the normal equations

$$
\begin{gathered}
X \beta=\mathbf{y} \Rightarrow X^{\top} X \beta=X^{\top} \mathbf{y} \\
X^{\top} X=\left(\begin{array}{ll}
3.00 & 5.13 \\
5.13 & 8.84
\end{array}\right) \quad X^{\top} \mathbf{y}=\binom{194.00}{341.29} \quad \hat{\beta}=\left(X^{\top} X\right)^{-1} X^{\top} \mathbf{y}=\binom{-173.14}{137.90} \\
\text { Weight }=-173.39+139.21 \text { Height }
\end{gathered}
$$

## Least-squares linear regression

## Example



MATLAB:

$$
\begin{aligned}
& \mathrm{X}=[11.70 ; 11.53 ; 11.90] ; \\
& \mathrm{y}=[57 ; 43 ; 94] ; \\
& \text { beta=inv }(\mathrm{X}, * \mathrm{X}) * \mathrm{X}{ }^{\prime} * \mathrm{y} \\
& \mathrm{x}=1.5: 0.01: 2.00 ; \\
& \text { yp=beta(1)+beta (2) *x; } \\
& \text { plot }\left(\mathrm{x}, \mathrm{yp}, \mathrm{X}(:, 1), \mathrm{y}, \mathrm{O}^{\prime}\right) \\
& \text { xlabel('Height (m)') } \\
& \text { ylabel('Weight (kg)') }
\end{aligned}
$$

## Least-squares linear regression

## The general linear model

The linear model is not restricted to straight lines. We can use it to fit any kind of curves:

$$
y=\beta_{0} f_{0}(x)+\beta_{1} f_{1}(x)+\beta_{2} f_{2}(x)+\ldots
$$

Fitting a parabola

$$
\begin{aligned}
f_{0}(x)=1 \\
f_{1}(x)=x \\
f_{2}(x)=x^{2}
\end{aligned} \Rightarrow \quad \begin{aligned}
& y_{1}=f_{0}\left(x_{1}\right)+\beta_{1} f_{1}\left(x_{1}\right)+\beta_{2} f_{2}\left(x_{1}\right) \\
& y_{2}=f_{0}\left(x_{2}\right)+\beta_{1} f_{1}\left(x_{2}\right)+\beta_{2} f_{2}\left(x_{2}\right) \\
& \ldots \\
& y_{n}=f_{0}\left(x_{n}\right)+\beta_{1} f_{1}\left(x_{n}\right)+\beta_{2} f_{2}\left(x_{n}\right) \\
&\left(\begin{array}{c}
y_{1} \\
y_{2} \\
\ldots \\
y_{n}
\end{array}\right)=\left(\begin{array}{ccc}
1 & x_{1} & x_{1}^{2} \\
1 & x_{2} & x_{2}^{2} \\
\ldots & \ldots & \ldots \\
1 & x_{n} & x_{n}^{2}
\end{array}\right)\left(\begin{array}{c}
\beta_{0} \\
\beta_{1} \\
\beta_{2}
\end{array}\right)+\left(\begin{array}{c}
\epsilon_{1} \\
\epsilon_{2} \\
\ldots \\
\epsilon_{n}
\end{array}\right) \quad \Rightarrow \mathbf{y}=X \beta+\epsilon
\end{aligned}
$$

## Least-squares linear regression

## Fitting a parabola

In this example they model the deformation of the wall of the zebra fish embryo as a function of strain.

Z. Lua, P. C.Y. Chen, H. Luo, J. Nam, R. Ge, W. Lin. Models of maximum stress and strain of zebrafish embryos under indentation. J. Biomechanics 42 (5): 620-625 (2009)

## Least-squares linear regression

## Multivariate linear regression

The linear model is not restricted to one variable. By fitting several variables we may fit surfaces and hypersurfaces

$$
y=\beta_{0} f_{0}\left(x_{1}, x_{2}\right)+\beta_{1} f_{1}\left(x_{1}, x_{2}\right)+\beta_{2} f_{2}\left(x_{1}, x_{2}\right)+\ldots
$$

## Fitting a parabolic surface

$$
\begin{aligned}
& f_{0}\left(x_{1}, x_{2}\right)=1 \\
& f_{1}\left(x_{1}, x_{2}\right)=x_{1} \\
& f_{2}\left(x_{1}, x_{2}\right)=x_{2} \\
& \Rightarrow X=\left(\begin{array}{cccccc}
1 & x_{11} & x_{12} & x_{11}^{2} & x_{12}^{2} & x_{11} x_{12} \\
1 & x_{21} & x_{22} & x_{21}^{2} & x_{22}^{2} & x_{21} x_{22} \\
\ldots & \ldots & \ldots & \ldots & \ldots & \ldots \\
1 & x_{n 1} & x_{n 2} & x_{n 1}^{2} & x_{n 2}^{2} & x_{n 1} x_{n 2}
\end{array}\right) \\
& f_{5}\left(x_{1}, x_{2}\right)=x_{1} x_{2}
\end{aligned}
$$

## Least-squares linear regression

## Fitting a parabolic surface

In this example they model the shape of cornea using videokeratoscopic images.


## Exercises

## Exercises

From Lay (3rd ed.), Chapter 6, Section 6:

- 6.6.1
- 6.6 .5
- 6.6 .9
- 6.6.12 (computer)


## Outline

(7) Orthogonality and least squares

- Inner product, length and orthogonality (a)
- Orthogonal sets, bases and matrices (a)
- Orthogonal projections (b)
- Gram-Schmidt orthogonalization (b)
- Least squares (c)
- Least-squares linear regression (c)
- Inner product spaces (d)
- Applications of inner product spaces (d)


## Inner product spaces



## Inner product spaces

## Definition 7.1 (Inner product)

An inner product in a vector space $V$ is a function that assigns a real number to every pair of vectors $\mathbf{u}$ and $\mathbf{v},\langle\mathbf{u}, \mathbf{v}\rangle$ and that satisfies the following axioms for all $\mathbf{u}, \mathbf{v}, \mathbf{w} \in V$ and all scalars $c$ :
(1) $\langle\mathbf{u}, \mathbf{v}\rangle=\langle\mathbf{v}, \mathbf{u}\rangle$
(2) $\langle\mathbf{u}+\mathbf{v}, \mathbf{w}\rangle=\langle\mathbf{u}, \mathbf{w}\rangle+\langle\mathbf{v}, \mathbf{w}\rangle$
(3) $\langle c \mathbf{u}, \mathbf{v}\rangle=c\langle\mathbf{u}, \mathbf{v}\rangle$
(4) $\langle\mathbf{u}, \mathbf{u}\rangle \geq 0$ and $\langle\mathbf{u}, \mathbf{u}\rangle=0$ iff $\mathbf{u}=\mathbf{0}$.

## Example

For instance in Weighted Least Squares (WLS) we may use an inner product in $\mathbb{R}^{2}$ defined as:

$$
\langle\mathbf{u}, \mathbf{v}\rangle=4 u_{1} v_{1}+5 u_{2} v_{2}
$$

In this way we give less weight to distances in the first component with respect to distances in the second component.

## Inner product spaces

Now we have to prove that this function is effectively an inner product:
(1) $\langle\mathbf{u}, \mathbf{v}\rangle=\langle\mathbf{v}, \mathbf{u}\rangle$

$$
\begin{aligned}
\langle\mathbf{u}, \mathbf{v}\rangle & =4 u_{1} v_{1}+5 u_{2} v_{2} & & \text { [by definition] } \\
& =4 v_{1} u_{1}+5 v_{2} u_{2} & & \text { [commutativity of scalar multiplication] } \\
& =\langle\mathbf{v}, \mathbf{u}\rangle & & \text { [by definition] }
\end{aligned}
$$

(2) $\langle\mathbf{u}+\mathbf{v}, \mathbf{w}\rangle=\langle\mathbf{u}, \mathbf{w}\rangle+\langle\mathbf{v}, \mathbf{w}\rangle$

$$
\begin{aligned}
\langle\mathbf{u}+\mathbf{v}, \mathbf{w}\rangle & =4\left(u_{1}+v_{1}\right) w_{1}+5\left(u_{2}+v_{2}\right) w_{2} \\
& =4 u_{1} w_{1}+4 v_{1} w_{1}+5 u_{2} w_{2}+5 v_{2} w_{2} \\
& =4 u_{1} w_{1}+5 u_{2} w_{2}+4 v_{1} w_{1}+5 v_{2} w_{2} \\
& =\langle\mathbf{u}, \mathbf{w}\rangle+\langle\mathbf{v}, \mathbf{w}\rangle
\end{aligned}
$$

[by definition]
[distributivity of scalar] [multiplication/addition] [commutativity] [of scalar addition] [by definition]

## Inner product spaces

(0) $\langle c \mathbf{u}, \mathbf{v}\rangle=c\langle\mathbf{u}, \mathbf{v}\rangle$

$$
\begin{aligned}
\langle c \mathbf{u}, \mathbf{v}\rangle & =4 c u_{1} v_{1}+5 c u_{2} v_{2} & & \text { [by definition] } \\
& =c 4 v_{1} u_{1}+c 5 v_{2} u_{2} & & \text { [commutativity of scalar multiplication] } \\
& =c\left(4 v_{1} u_{1}+5 v_{2} u_{2}\right) & & \text { [distributivity of scalar multiplication] } \\
& =c\langle\mathbf{u}, \mathbf{v}\rangle & & \text { [by definition] }
\end{aligned}
$$

(c) $\langle\mathbf{u}, \mathbf{u}\rangle \geq 0$ and $\langle\mathbf{u}, \mathbf{u}\rangle=0$ iff $\mathbf{u}=\mathbf{0}$.
(1) $\langle\mathbf{u}, \mathbf{u}\rangle \geq 0$

$$
\langle\mathbf{u}, \mathbf{u}\rangle=4 u_{1}^{2}+5 u_{2}^{2} \quad[\text { by definition }]
$$

which is obviously larger than 0 .
(2) $\langle\mathbf{u}, \mathbf{u}\rangle=0$ iff $\mathbf{u}=\mathbf{0}$.

$$
\langle\mathbf{u}, \mathbf{u}\rangle=0 \Leftrightarrow 4 u_{1}^{2}+5 u_{2}^{2}=0 \Leftrightarrow u_{1}=u_{2}=0
$$

## Inner product spaces

## Example

Consider two vectors $p$ and $q$ the vector space of polynomials of degree $n\left(\mathbb{P}_{n}\right)$. Let $t_{0}, t_{1}, \ldots, t_{n}$ be $n$ distinct real numbers and $K$ any scalar. The inner product between $p$ and $q$ is defined as

$$
\langle p, q\rangle=K\left(p\left(t_{0}\right) q\left(t_{0}\right)+p\left(t_{1}\right) q\left(t_{1}\right)+\ldots+p\left(t_{n}\right) q\left(t_{n}\right)\right)
$$

Axioms 1-3 are easy to check. Let's prove Axiom 4

- $\langle p, p\rangle \geq 0$ and $\langle p, p\rangle=0$ iff $p=0$.
- $\langle p, p\rangle \geq 0$

$$
\langle p, p\rangle=K\left(p^{2}\left(t_{0}\right)+p^{2}\left(t_{1}\right)+\ldots+p^{2}\left(t_{n}\right)\right) \quad[\text { by definition }]
$$

which is obviously larger than 0 .
(3) $\langle p, p\rangle=0$ iff $p=0$.

$$
\begin{gathered}
\langle p, p\rangle=0 \Leftrightarrow K\left(p^{2}\left(t_{0}\right)+p^{2}\left(t_{1}\right)+\ldots+p^{2}\left(t_{n}\right)\right) \Leftrightarrow \\
p\left(t_{0}\right)=p\left(t_{1}\right)=\ldots=p\left(t_{n}\right)=0
\end{gathered}
$$

But $p$ is a polynomial of degree $n$ so, at most, it can have $n$ zeros. However, the previous condition requires the polynomial to vanish at $n+1$ points. This is impossible unless $p=0$.

## Inner product spaces

## Example

Consider two vectors $p$ and $q$ the vector space of polynomials of degree $n\left(\mathbb{P}_{n}\right)$. Assume that we regularly space the $n+1$ points in the interval $[-1,1]$

and set $K=\Delta T$, then the inner product between the two polynomials becomes

$$
\langle p, q\rangle=\left(p\left(t_{0}\right) q\left(t_{0}\right)+p\left(t_{1}\right) q\left(t_{1}\right)+\ldots+p\left(t_{n}\right) q\left(t_{n}\right)\right) \Delta T=\sum_{i=0}^{n} p\left(t_{i}\right) q\left(t_{i}\right) \Delta T
$$

Making $\Delta T$ tend to 0 , the inner product becomes

$$
\langle p, q\rangle=\int_{-1}^{1} p(t) q(t) d t
$$

## Inner product spaces

Legendre polynomials are orthogonal polynomials in the interval $[-1,1]$

$$
\begin{aligned}
& P_{0}(x)=1 \\
& P_{1}(x)=x \\
& P_{2}(x)=\frac{1}{2}\left(3 x^{2}-1\right) \\
& P_{3}(x)=\frac{1}{2}\left(5 x^{3}-3 x\right) \\
& P_{4}(x)=\frac{1}{8}\left(35 x^{4}-30 x^{2}+3\right) \\
& P_{5}(x)=\frac{1}{8}\left(63 x^{5}-70 x^{3}+15 x\right) \\
& P_{6}(x)=\frac{1}{16}\left(231 x^{6}-315 x^{4}+105 x^{2}-5\right) \\
& P_{7}(x)=\frac{1}{16}\left(429 x^{7}-693 x^{5}+315 x^{3}-35 x\right)
\end{aligned}
$$



Legendre polynomials are very useful for regression of high-order polynomials as shown in next slide.

## Inner product spaces

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4.

## Length, distance and orthogonality

Length, distance and orthogonality
The length of a vector $\mathbf{u}$ in an inner product space is defined in the standard way

$$
\|\mathbf{u}\|=\sqrt{\langle\mathbf{u}, \mathbf{u}\rangle}
$$

Similarly, the distance between two vectors $\mathbf{u}$ and $\mathbf{v}$ is defined as

$$
d(\mathbf{u}, \mathbf{v})=\|\mathbf{u}-\mathbf{v}\|
$$

Finally, two vectors $\mathbf{u}$ and $\mathbf{v}$ are said to be orthogonal iff

$$
\langle\mathbf{u}, \mathbf{v}\rangle=0
$$

## Length, distance and orthogonality

## Example

In the vector space of polynomials in the interval $[0,1], \mathbb{P}[0,1]$, let's define the inner product

$$
\langle p, q\rangle=\int_{0}^{1} p(t) q(t) d t
$$

What is the length of the vector $p(t)=3 t^{2}$ ?
Solution

$$
\begin{aligned}
\|p\| & =\sqrt{\langle p, p\rangle}=\sqrt{\int_{0}^{1} p^{2}(t) d t}=\sqrt{\int_{0}^{1}\left(3 t^{2}\right)^{2} d t}=\sqrt{\int_{0}^{1} 9 t^{4} d t} \\
& =\sqrt{\left.9 \frac{t^{5}}{5}\right|_{0} ^{1}}=\sqrt{9\left(\frac{1}{5}-0\right)}=\frac{3}{\sqrt{5}}
\end{aligned}
$$

## Gram-Schmidt orthogonalization

## Example

Gram-Schmidt is applied in the standard way. For instance, find an orthogonal basis of $\mathbb{P}_{2}[-1,1]$. A basis that spans this space is

$$
\left\{1, t, t^{2}\right\}
$$

Let's orthogonalize it

$$
\begin{aligned}
p_{0}(t) & =1 \\
p_{1}(t) & =t-\frac{\left\langle t, p_{0}(t)\right\rangle}{\left\|p_{0}\right\|^{2}} p_{0}(t)=t-\frac{\int_{-1}^{1} t d t}{\int_{-1}^{1} d t} 1=t-\frac{0}{2} 1=t \\
p_{2}(t) & =t^{2}-\frac{\left\langle t^{2}, p_{0}(t)\right\rangle}{\left\|p_{0}\right\|^{2}} p_{0}(t)-\frac{\left\langle t^{2}, p_{1}(t)\right\rangle}{\left\|p_{1}\right\|^{2}} p_{1}(t) \\
& =t^{2}-\frac{\int_{-1}^{\| t^{2} d t}}{\int_{-1}^{1} d t}-\frac{\int_{-1}^{1} t^{2} t d t}{\int_{-1}^{1} t^{2} d t} t=t^{2}-\frac{2}{3}=t^{2}-\frac{1}{3}
\end{aligned}
$$

In Slide 97 we proposed the Legendre polynomial of degree 2 to be $P_{2}(t)=\frac{1}{2}\left(3 t^{2}-1\right)$, we can easily show that $P_{2}(t)=\frac{3}{2} p_{2}(t)$. Consequently, if $p_{2}(t)$ is orthogonal to $p_{0}(t)$ and $p_{1}(t)$ so is $P_{2}(t)$.

## Best approximation

## Example

What is the best approximation in $\mathbb{P}_{2}[-1,1]$ of $p(t)=t^{3}$ ?

## Solution

We know the answer is the orthogonal projection of $p(t)$ onto $\mathbb{P}_{2}[-1,1]$. An orthogonal basis of $\mathbb{P}_{2}[-1,1]$ is $\left\{1, t, t^{2}-\frac{1}{3}\right\}$. Therefore, this projection can be calculated as

$$
\hat{p}(t)=\operatorname{Proj}_{\mathbb{P}_{2}[-1,1]}\{p(t)\}=\frac{\left\langle p, p_{0}\right\rangle}{\left\|p_{0}\right\|^{2}} p_{0}(t)+\frac{\left\langle p, p_{1}\right\rangle}{\left\|p_{1}\right\|^{2}} p_{1}(t)+\frac{\left\langle p, p_{2}\right\rangle}{\left\|p_{2}\right\|^{2}} p_{2}(t)
$$

Let's perform these calculations:

\[

\]

## Best approximation



## Best approximation

## Example

In this example we exploited the best approximation property of orthogonal wavelets to speed-up and make more robust angular alignments of projections in 3D Electron Microscopy.


[^1]
## Pythagorean theorem

## Theorem 7.1 (Pythagorean theorem)

Given any vector $\mathbf{v}$ in an inner product space $V$ and a subspace of it $W \subseteq V$ we have

$$
\|\mathbf{v}\|^{2}=\left\|\operatorname{Proj}_{W}\{\mathbf{v}\}\right\|^{2}+\left\|\mathbf{v}-\operatorname{Proj}_{W}\{\mathbf{v}\}\right\|^{2}
$$



## FIGURE 2

The hypotenuse is the longest side.

## The Cauchy-Schwarz inequality

## Theorem 7.2 (The Cauchy-Schwarz inequality)

For all $\mathbf{u}, \mathbf{v} \in V$ it is verified

$$
|\langle\mathbf{u}, \mathbf{v}\rangle| \leq\|\mathbf{u}\|\|\mathbf{v}\|
$$

Proof
If $\mathbf{u}=\mathbf{0}$, then

$$
|\langle\mathbf{0}, \mathbf{v}\rangle|=0 \quad \text { and } \quad\|\mathbf{0}\|\|\mathbf{v}\|=0\|\mathbf{v}\|=0
$$

So the inequality becomes an equality. If $\mathbf{u} \neq \mathbf{0}$, then consider $W=\operatorname{Span}\{\mathbf{u}\}$ and

$$
\left\|\operatorname{Proj}_{W}\{\mathbf{v}\}\right\|=\left\|\frac{\langle\mathbf{v}, \mathbf{u}\rangle}{\|\mathbf{u}\|^{2}} \mathbf{u}\right\|=\frac{|\langle\mathbf{v}, \mathbf{u}\rangle|}{\|\mathbf{u}\|^{2}}\|\mathbf{u}\|=\frac{|\langle\mathbf{v}, \mathbf{u}\rangle|}{\|\mathbf{u}\|}
$$

But by the Pythagorean Theorem (Theorem 7.1) we have $\left\|\operatorname{Proj}_{w}\{\mathbf{v}\}\right\| \leq\|\mathbf{v}\|$. Consequently,

$$
\frac{|\langle\mathbf{v}, \mathbf{u}\rangle|}{\|\mathbf{u}\|} \leq\|\mathbf{v}\| \Rightarrow|\langle\mathbf{v}, \mathbf{u}\rangle| \leq\|\mathbf{u}\|\|\mathbf{v}\| \text { (q.e.d.) }
$$

## The Triangle inequality

## Theorem 7.3 (The Triangle inequality)

For all $\mathbf{u}, \mathbf{v} \in V$ it is verified

$$
\|\mathbf{u}+\mathbf{v}\| \leq\|\mathbf{u}\|+\|\mathbf{v}\|
$$

Proof

$$
\begin{aligned}
\|\mathbf{u}+\mathbf{v}\|^{2} & =\langle\mathbf{u}+\mathbf{v}, \mathbf{u}+\mathbf{v}\rangle & & \text { [By definition] } \\
& =\langle\mathbf{u}, \mathbf{u}\rangle+\langle\mathbf{v}, \mathbf{v}\rangle+2\langle\mathbf{u}, \mathbf{v}\rangle & & \text { [Properties of inner product] } \\
& \leq\|\mathbf{u}\|^{2}+\|\mathbf{v}\|^{2}+2|\langle\mathbf{u}, \mathbf{v}\rangle| & & \langle\mathbf{u}, \mathbf{v}\rangle \leq|\langle\mathbf{u}, \mathbf{v}\rangle| \\
& \leq\|\mathbf{u}\|^{2}+\|\mathbf{v}\|^{2}+2\|\mathbf{u}\|\|\mathbf{v}\| & & \text { Cauchy-Schwarz } \\
& =(\|\mathbf{u}\|+\|\mathbf{v}\|)^{2} & & \\
& \Rightarrow \mathbf{u}+\mathbf{v} \| & \leq\|\mathbf{u}\|+\|\mathbf{v}\| &
\end{aligned}
$$

(q.e.d.)

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 6, Section 7:

- 6.7.1
- 6.7.13
- 6.7.16
- 6.7.18


## Outline

(7) Orthogonality and least squares

- Inner product, length and orthogonality (a)
- Orthogonal sets, bases and matrices (a)
- Orthogonal projections (b)
- Gram-Schmidt orthogonalization (b)
- Least squares (c)
- Least-squares linear regression (c)
- Inner product spaces (d)
- Applications of inner product spaces (d)


## Weighted Least Squares

## Weighted Least Squares

Let us assume we have a table of collected data and we want to fit a least squares model. However, we want to give more importance to some observations because we are more confident about them or they are more important. We encode the importance as a weight value (the larger the weight, the more importance the observation has)

| $\mathbf{X}$ | $\mathbf{Y}$ | $\mathbf{W}$ |
| :---: | :---: | :---: |
| $x_{1}$ | $y_{1}$ | $w_{1}$ |
| $x_{2}$ | $y_{2}$ | $w_{2}$ |
| $x_{3}$ | $y_{3}$ | $w_{3}$ |
| $\cdots$ | $\cdots$ | $\cdots$ |

Let us call $\hat{y}_{j}$ the prediction of the model for the $j$-th observation and $\epsilon_{j}$ the committed error

$$
y_{j}=\hat{y}_{j}+\epsilon_{j}
$$

## Weighted Least Squares

The goal is now to minimize the weighted sum of square errors

$$
\sum_{j=1}^{n}\left(w_{j} \epsilon_{j}\right)^{2}=\sum_{j=1}^{n}\left(w_{j}\left(y_{j}-\hat{y}_{j}\right)\right)^{2}=\sum_{j=1}^{n}\left(w_{j} y_{j}-w_{j} \hat{y}_{j}\right)^{2}
$$

Let us collect all observed values into a vector $\mathbf{y}$ and do analogously with the predictions $\hat{\mathbf{y}}$. Let us define the diagonal matrix

$$
W=\left(\begin{array}{ccccc}
w_{1} & 0 & 0 & \ldots & 0 \\
0 & w_{2} & 0 & \ldots & 0 \\
0 & 0 & w_{3} & \ldots & 0 \\
\ldots & \ldots & \ldots & \ldots & \ldots \\
0 & 0 & 0 & \ldots & w_{n}
\end{array}\right)
$$

Then, the previous objective function becomes

$$
\sum_{j=1}^{n}\left(w_{j} y_{j}-w_{j} \hat{y}_{j}\right)^{2}=\|W \mathbf{y}-W \hat{\mathbf{y}}\|^{2}
$$

## Weighted Least Squares

Now, suppose that $\hat{\mathbf{y}}$ is calculated from the columns of a matrix $A$, that is, $\hat{\mathbf{y}}=A \mathbf{x}$. The objective function becomes

$$
\sum_{j=1}^{n}\left(w_{j} y_{j}-w_{j} \hat{y}_{j}\right)^{2}=\|W \mathbf{y}-W A \mathbf{x}\|^{2}
$$

The minimum of this objective function is attained for $\hat{\mathrm{x}}$ that is the least-squares solution of the equation system

$$
W A \mathbf{x}=W \mathbf{y}
$$

The normal equations of the problem are

$$
(W A)^{T} W A \mathbf{x}=(W A)^{T} W \mathbf{y}
$$

## Weighted Least Squares

## Example

In this work they used Weighted Least Squares to calibrate a digital system to measure maximum respiratory pressures.

J.L. Ferreira, F.H. Vasconcelos, C.J. Tierra-Criollo. A Case Study of Applying Weighted Least Squares to Calibrate a Digital Maximum Respiratory Pressures Measuring System. Applied Biomedical Engineering, Chapter 18 (2011)

## Fourier Series

## Example

Fourier tools are, maybe, the most widespread tool to analyze signals and its frequency components. Fourier decomposition states that any signal can be obtained by summing sine waves of different amplitude, phase and frequency.


## Fourier Series

## Theorem 8.1

Consider the vector space of continuous functions in the interval $[0,2 \pi], C[0,2 \pi]$. The set

$$
S=\{1, \cos (t), \sin (t), \cos (2 t), \sin (2 t), \ldots, \cos (N t), \sin (N t)\}
$$

is orthogonal with respect to the inner product defined as

$$
\langle f(t), g(t)\rangle=\int_{0}^{2 \pi} f(t) g(t) d t
$$

Proof

$$
\begin{aligned}
\langle\cos (n t), \cos (m t)\rangle & =\int_{0}^{2 \pi} \cos (n t) \cos (m t) d t \\
& =\int_{0}^{2 \pi} \frac{1}{2}(\cos ((n+m) t)+\cos ((n-m) t)) d t \\
& =\left.\frac{1}{2}\left(\frac{\sin ((n+m) t)}{n+m}+\frac{\sin ((n-m) t)}{n-m}\right)\right|_{0} ^{2 * \pi} \\
& =0
\end{aligned}
$$

where we have used $\cos (A) \cos (B)=\frac{1}{2}(\cos (A+B)+\cos (A-B))$.

## Fourier Series

Analogously we could prove that

$$
\begin{aligned}
\langle\cos (n t), \sin (m t)\rangle & =0 \\
\langle\cos (n t), 1\rangle & =0 \\
\langle\sin (n t), 1\rangle & =0 \\
\|\cos (n t)\|^{2} & =\pi \\
\|\sin (n t)\|^{2} & =\pi \\
\|1\|^{2} & =2 \pi
\end{aligned}
$$

## Fourier Series

## Theorem 8.2 (Fourier series)

Given any function $f(t) \in C[0,2 \pi], f(t)$ can be approximated as closely as desired by a sum of the form simply by orthogonally projecting it onto $W=\operatorname{Span}\{S\}$

$$
f(t) \approx \operatorname{Proj}_{W}\{f(t)\}=\frac{\langle f(t), 1\rangle}{\|1\|^{2}}+\sum_{n=1}^{N}\left(\frac{\langle f(t), \cos (n t)\rangle}{\|\cos (n t)\|^{2}} \cos (n t)+\frac{\langle f(t), \sin (n t)\rangle}{\|\sin (n t)\|^{2}} \sin (n t)\right)
$$




## Fourier Series

## Example

In this work we used Fourier space to simulate and to align electron microscopy images

S. Jonic, C.O.S.Sorzano, P. Thévenaz, C. EI-Bez, S. De Carlo, M. Unser. Spline-Based image-to-volume registration for three-dimensional electron microscopy. Ultramicroscopy, 103: 303-317 (2005)

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 6, Section 8:

- 6.8.1
- 6.8.6
- 6.8 .8
- 6.8.11


## Outline

(7) Orthogonality and least squares

- Inner product, length and orthogonality (a)
- Orthogonal sets, bases and matrices (a)
- Orthogonal projections (b)
- Gram-Schmidt orthogonalization (b)
- Least squares (c)
- Least-squares linear regression (c)
- Inner product spaces (d)
- Applications of inner product spaces (d)


# Chapter 8. Symmetric matrices and quadratic forms 

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## Outline

(8) Symmetric matrices and quadratic forms

- Diagonalization of symmetric matrices (a)
- Quadratic forms (b)
- Constrained optimization (b)
- Singular Value Decomposition (SVD) (c)


## References


D. Lay. Linear algebra and its applications (3rd ed). Pearson (2006). Chapter 7.

## Applications

In this example of particle picking in Single Particles, one of the features we analyze is the autocorrelation function at different subbands. The autocorrelation is a symmetric matrix.

V. Abrishami, A. Zaldívar-Peraza, J.M. de la Rosa-Trevín, J. Vargas, J. Otón, R. Marabini, Y. Shkolnisky, J.M. Carazo, C.O.S. Sorzano. A pattern matching approach to the automatic selection of particles from low-contrast electron micrographs. Bioinformatics (2013)

## Applications

In one of the steps, we construct a basis that spans the set of rotations of the particle template. For doing so, perform a Principal Component Analysis that diagonalizes the covariance matrix (which is again a symmetric matrix).

(a)

V. Abrishami, A. Zaldívar-Peraza, J.M. de la Rosa-Trevín, J. Vargas, J. Otón, R. Marabini, Y. Shkolnisky, J.M. Carazo, C.O.S. Sorzano. A pattern matching approach to the automatic selection of particles from low-contrast electron micrographs. Bioinformatics (2013)

## Outline

8 Symmetric matrices and quadratic forms

- Diagonalization of symmetric matrices (a)
- Quadratic forms (b)
- Constrained optimization (b)
- Singular Value Decomposition (SVD) (c)


## Diagonalization of symmetric matrices

## Definition 1.1 (Symmetric matrix)

$A \in \mathcal{M}_{n \times n}$ is a symmetric matrix iff $A=A^{T}$.

## Example

The following two matrices are symmetric

$$
\left(\begin{array}{cc}
1 & 0 \\
0 & -3
\end{array}\right) \quad\left(\begin{array}{ccc}
0 & -1 & 0 \\
-1 & 5 & 8 \\
0 & 8 & -7
\end{array}\right)
$$

## Example

Let's diagonalize the matrix $A=\left(\begin{array}{ccc}6 & -2 & -1 \\ -2 & 6 & -1 \\ -1 & -1 & 5\end{array}\right)$ The characteristic equation is

$$
|A-\lambda I|=0=-\lambda^{3}+17 \lambda^{2}-90 \lambda+144=-(\lambda-8)(\lambda-6)(\lambda-3)
$$

## Diagonalization of symmetric matrices

The associated eigenvectors are

$$
\begin{array}{ll}
\lambda=8 & \mathbf{v}_{1}=(-1,1,0) \rightarrow \mathbf{u}_{1}=\left(-\frac{1}{\sqrt{2}}, \frac{1}{\sqrt{2}}, 0\right) \\
\lambda=6 & \mathbf{v}_{2}=(-1,-1,2) \rightarrow \mathbf{u}_{2}=\left(-\frac{1}{\sqrt{6}},-\frac{1}{\sqrt{6}}, \frac{2}{\sqrt{6}}\right) \\
\lambda=3 & \mathbf{v}_{3}=(1,1,1) \rightarrow \mathbf{u}_{3}=\left(\frac{1}{\sqrt{3}}, \frac{1}{\sqrt{3}}, \frac{1}{\sqrt{3}}\right)
\end{array}
$$

The $\mathbf{v}$ vectors constitute an orthogonal basis of $\mathbb{R}^{3}$ and after normalizing them $\left(\mathbf{u}_{i}=\frac{\mathbf{v}_{i}}{\left\|\mathbf{v}_{i}\right\|}\right)$, we have an orthonormal basis Thus, we can factorize $A$ as $A=P D P^{-1}$ with

$$
P=\left(\begin{array}{ccc}
-\frac{1}{\sqrt{2}} & -\frac{1}{\sqrt{6}} & \frac{1}{\sqrt{3}} \\
\frac{1}{\sqrt{2}} & -\frac{1}{\sqrt{6}} & \frac{1}{\sqrt{3}} \\
0 & \frac{2}{\sqrt{6}} & \frac{1}{\sqrt{3}}
\end{array}\right) \quad D=\left(\begin{array}{lll}
8 & 0 & 0 \\
0 & 6 & 0 \\
0 & 0 & 3
\end{array}\right)
$$

Exploiting the fact that $P$ is orthonormal, then $P^{-1}=P^{T}$ and $A=P D P^{T}$.

## Diagonalization of symmetric matrices

## Theorem 1.1

If $A$ is symmetric, then any two eigenvectors from different eigenspaces are orthogonal.
Proof
Let $\mathbf{v}_{1}$ and $\mathbf{v}_{2}$ be two eigenvectors from two different eigenvalues $\lambda_{1}$ and $\lambda_{2}$. Let's show that $\mathbf{v}_{1} \cdot \mathbf{v}_{2}=0$

$$
\begin{aligned}
\lambda_{1}\left(\mathbf{v}_{1} \cdot \mathbf{v}_{2}\right) & =\left(\lambda_{1} \mathbf{v}_{1}\right)^{T} \mathbf{v}_{2} & & \text { [By definition] } \\
& =\left(A \mathbf{v}_{1}\right)^{T} \mathbf{v}_{2} & & \text { [Definition of eigenvector] } \\
& =\mathbf{v}_{1}^{T} A^{T} \mathbf{v}_{2} & & \text { [Transpose of product] } \\
& =\mathbf{v}_{1}^{T}\left(A \mathbf{v}_{2}\right) & & \text { [A is symmetric] } \\
& =\mathbf{v}_{1}^{T}\left(\lambda_{2} \mathbf{v}_{2}\right) & & \text { [Definition of eigenvector] } \\
& =\lambda_{2}\left(\mathbf{v}_{1} \cdot \mathbf{v}_{2}\right) & & \text { [By definition] }
\end{aligned}
$$

Hence $\left(\lambda_{1}-\lambda_{2}\right)\left(\mathbf{v}_{1} \cdot \mathbf{v}_{2}\right)=0$ but $\lambda_{1}-\lambda_{2} \neq 0$ because the two eigenvalues are different. Consequently, $\mathbf{v}_{1} \cdot \mathbf{v}_{2}=0$ (q.e.d.)

## Diagonalization of symmetric matrices

## Definition 1.2 (Orthogonal diagonalization)

$A$ is orthogonally diagonalizable iff $A=P D P^{\top}$ being $P$ an orthogonal (i.e., $P^{-1}=P^{T}$ ).

## Theorem 1.2

$A$ is orthogonally diagonalizable iff $A$ is symmetric.
Proof orthogonally diagonalizable $\Rightarrow$ symmetric
Let us assume that $A=P D P^{T}$, then

$$
A^{T}=\left(P D P^{T}\right)^{T}=\left(P^{T}\right)^{T} D^{T} P^{T}=P D^{T} P^{T}=P D P^{T}=A
$$

Proof orthogonally diagonalizable $\Leftarrow$ symmetric
We omit this proof since it is more difficult.

## Diagonalization of symmetric matrices

## Example

Let's orthogonally diagonalize $A=\left(\begin{array}{ccc}3 & -2 & 4 \\ -2 & 6 & 2 \\ 4 & 2 & 3\end{array}\right)$.

## Solution

The characteristic equation is

$$
|A-\lambda I|=0=-\lambda^{3}+12 \lambda^{2}-21 \lambda-98=-(\lambda-7)^{2}(\lambda+2)
$$

Its associated eigenvectors are

$$
\begin{array}{ll}
\lambda=7 & \mathbf{v}_{1}=(1,0,1) \rightarrow \mathbf{u}_{1}=\left(\frac{1}{\sqrt{2}}, 0, \frac{1}{\sqrt{2}}\right) \\
& \mathbf{v}_{2}=\left(-\frac{1}{2}, 1,2\right) \rightarrow \mathbf{u}_{2}=\left(-\frac{1}{\sqrt{5}}, \frac{2}{\sqrt{5}}, 0\right) \\
\lambda=-2 & \mathbf{v}_{3}=\left(-1,-\frac{1}{2}, 1\right) \rightarrow \mathbf{u}_{3}=\left(-\frac{2}{3},-\frac{1}{3}, \frac{2}{3}\right)
\end{array}
$$

$\mathbf{u}_{1}$ and $\mathbf{u}_{2}$ are unitary and linearly independent, but they are not orthogonal. $\mathbf{u}_{3}$ is orthogonal to the other two vectors because it belongs to a different eigenspace (see Theorem 1.1).

## Diagonalization of symmetric matrices

We can orthogonalize $\mathbf{u}_{1}$ and $\mathbf{u}_{2}$ following the Gram-Schmidt procedure:

$$
\begin{aligned}
& \mathbf{w}_{1}=\mathbf{v}_{1}=\left(\frac{1}{\sqrt{2}}, 0, \frac{1}{\sqrt{2}}\right) \\
& \mathbf{w}_{2}^{\prime}=\mathbf{v}_{2}-\left\langle\mathbf{v}_{2}, \mathbf{w}_{1}\right\rangle \mathbf{w}_{1}=\left(-\frac{1}{\sqrt{5}}, \frac{2}{\sqrt{5}}, 0\right)-\left(-\frac{1}{\sqrt{10}}\right)\left(\frac{1}{\sqrt{2}}, 0, \frac{1}{\sqrt{2}}\right)=\left(-\frac{1}{2 \sqrt{5}}, \frac{2}{\sqrt{5}}, \frac{1}{2 \sqrt{5}}\right) \\
& \mathbf{w}_{2}=\frac{\mathbf{w}_{2}^{\prime}}{\left\|\mathbf{w}_{2}\right\|}=\left(-\frac{1}{3 \sqrt{2}}, \frac{2 \sqrt{2}}{3}, \frac{1}{3 \sqrt{2}}\right) \\
& \mathbf{w}_{3}=\mathbf{v}_{3}=\left(-\frac{2}{3},-\frac{1}{3}, \frac{2}{3}\right)
\end{aligned}
$$

So $A=P D P^{T}$ with

$$
P=\left(\begin{array}{ccc}
\frac{1}{\sqrt{2}} & -\frac{1}{3 \sqrt{2}} & -\frac{2}{3} \\
0 & \frac{2 \sqrt{2}}{3} & -\frac{1}{3} \\
\frac{1}{\sqrt{2}} & \frac{1}{3 \sqrt{2}} & \frac{2}{3}
\end{array}\right) \quad D=\left(\begin{array}{ccc}
7 & 0 & 0 \\
0 & 7 & 0 \\
0 & 0 & -2
\end{array}\right)
$$

## Diagonalization of symmetric matrices

## Definition 1.3 (Spectrum of a matrix)

The set of eigenvalues of a matrix is called the spectrum of that matrix.

## Theorem 1.3 (Spectral theorem for symmetric matrices)

An $n \times n$ symmetric matrix has the following properties:
(1) A has $n$ real eigenvalues (including multiplicities).
(2) The dimension of each eigenspace is the multiplicity of the corresponding eigenvalue as root of the characteristic equation.
(3) Eigenspaces corresponding to distinct eigenvalues are mutually orthogonal.
(4) $A$ is orthogonally diagonalizable.

## Diagonalization of symmetric matrices

## Definition 1.4 (Spectral decomposition of symmetric matrices)

Let $A=P D P^{\top}$ with $P=\left(\begin{array}{llll}\mathbf{u}_{1} & \mathbf{u}_{2} & \ldots & \mathbf{u}_{n}\end{array}\right)$. Then

$$
\begin{aligned}
A & =\left(\begin{array}{llll}
\mathbf{u}_{1} & \mathbf{u}_{2} & \ldots & \mathbf{u}_{n}
\end{array}\right)\left(\begin{array}{cccc}
\lambda_{1} & 0 & \ldots & 0 \\
0 & \lambda_{2} & \ldots & 0 \\
0 & 0 & \ldots & \lambda_{n}
\end{array}\right)\left(\begin{array}{c}
\mathbf{u}_{1}^{T} \\
\mathbf{u}_{2}^{T} \\
\ldots \\
\mathbf{u}_{n}^{T}
\end{array}\right) \\
& =\left(\begin{array}{llll}
\lambda_{1} \mathbf{u}_{1} & \lambda_{2} \mathbf{u}_{2} & \ldots & \lambda_{n} \mathbf{u}_{n}
\end{array}\right)\left(\begin{array}{c}
\mathbf{u}_{1}^{T} \\
\mathbf{u}_{2}^{T} \\
\ldots \\
\mathbf{u}_{n}^{T}
\end{array}\right) \\
& =\lambda_{1} \mathbf{u}_{1} \mathbf{u}_{1}^{T}+\lambda_{2} \mathbf{u}_{2} \mathbf{u}_{2}^{T}+\ldots+\lambda_{n} \mathbf{u}_{n} \mathbf{u}_{n}^{T}
\end{aligned}
$$

The latest equation is the spectral decomposition of $A$. Each one of the terms $\lambda_{i} \mathbf{u}_{i} \mathbf{u}_{i}^{T}$ is an $n \times n$ matrix of rank 1 (since all the columns are multiples of $\mathbf{u}_{i}$. Additionally, $\mathbf{u}_{i} \mathbf{u}_{i}^{T} \mathbf{x}$ is the orthogonal projection of any vector onto the subspace generated by $\mathbf{u}_{i}$.

## Diagonalization of symmetric matrices

## Example

Write the spectral decomposition of

$$
A=\left(\begin{array}{cc}
\frac{2}{\sqrt{5}} & -\frac{1}{\sqrt{5}} \\
\frac{1}{\sqrt{5}} & \frac{2}{\sqrt{5}}
\end{array}\right)\left(\begin{array}{ll}
8 & 0 \\
0 & 3
\end{array}\right)\left(\begin{array}{cc}
\frac{2}{\sqrt{5}} & \frac{1}{\sqrt{5}} \\
-\frac{1}{\sqrt{5}} & \frac{2}{\sqrt{5}}
\end{array}\right)
$$

## Solution

Consider $\mathbf{u}_{1}=\left(\frac{2}{\sqrt{5}}, \frac{1}{\sqrt{5}}\right)$ be the first column of $P$ and $\mathbf{u}_{2}=\left(-\frac{1}{\sqrt{5}}, \frac{2}{\sqrt{5}}\right)$. Then

$$
\mathbf{u}_{1} \mathbf{u}_{1}^{T}=\left(\begin{array}{cc}
\frac{4}{5} & \frac{2}{5} \\
\frac{2}{5} & \frac{1}{5}
\end{array}\right) \quad \mathbf{u}_{2} \mathbf{u}_{2}^{T}=\left(\begin{array}{cc}
\frac{1}{5} & -\frac{2}{5} \\
-\frac{2}{5} & \frac{4}{5}
\end{array}\right)
$$

The spectral decomposition is therefore

$$
A=\lambda_{1} \mathbf{u}_{1} \mathbf{u}_{1}^{T}+\lambda_{2} \mathbf{u}_{2} \mathbf{u}_{2}^{T}=8\left(\begin{array}{cc}
\frac{4}{5} & \frac{2}{5} \\
\frac{2}{5} & \frac{1}{5}
\end{array}\right)+3\left(\begin{array}{cc}
\frac{1}{5} & -\frac{2}{5} \\
-\frac{2}{5} & \frac{4}{5}
\end{array}\right)
$$

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 7, Section 1:

- 7.1.6
- 7.1.7
- 7.1.13
- 7.1.23
- 7.1.27
- 7.1.29
- 7.1.35


## Outline

(8) Symmetric matrices and quadratic forms

- Diagonalization of symmetric matrices (a)
- Quadratic forms (b)
- Constrained optimization (b)
- Singular Value Decomposition (SVD) (c)


## Quadratic forms

## Introduction

Most expressions appearing so far are linear: $A \mathbf{x},\langle\mathbf{w}, \mathbf{x}\rangle, \mathbf{x}^{T}$, that is, if we construct an operator $T(\mathbf{x})$ with them (e.g., $T(\mathbf{x})=A \mathbf{x}, T(\mathbf{x})=\langle\mathbf{w}, \mathbf{x}\rangle$, $T(\mathbf{x})=\mathbf{x}^{T}$ ), it meets

$$
T\left(a \mathbf{x}_{1}+b \mathbf{x}_{2}\right)=a T\left(\mathbf{x}_{1}\right)+b T\left(\mathbf{x}_{2}\right)
$$

However, there are nonlinear expressions like $\mathbf{x}^{T} \mathbf{x}$. Particularly, this one is said to be quadratic and they normally appear in applications of linear algebra to engineering (like optimization) and signal processing (like signal power). They also arise in physics (as potential and kinetic energy), differential geometry (as the normal curvature of surfaces) and statistics (as confidence ellipsoids).

## Quadratic forms

## Definition 2.1 (Quadratic forms)

A quadratic form in $\mathbb{R}^{n}$ is a function $Q(\mathbf{x}): \mathbb{R}^{n} \rightarrow \mathbb{R}$ that can be computed as

$$
Q(\mathbf{x})=\mathbf{x}^{T} A \mathbf{x}
$$

being $A \in \mathcal{M}_{n \times n}$ a symmetric matrix.

## Example

(1) $Q(\mathbf{x})=\mathbf{x}^{T} / \mathbf{x}=\left(\begin{array}{ll}x_{1} & x_{2}\end{array}\right)\left(\begin{array}{ll}1 & 0 \\ 0 & 1\end{array}\right)\binom{x_{1}}{x_{2}}=x_{1}^{2}+x_{2}^{2}$
(2) $Q(\mathbf{x})=\mathrm{x}^{T}\left(\begin{array}{ll}4 & 0 \\ 0 & 3\end{array}\right) \mathrm{x}=4 x_{1}^{2}+3 x_{2}^{2}$

- $Q(\mathbf{x})=\mathbf{x}^{T}\left(\begin{array}{cc}3 & -2 \\ -2 & 7\end{array}\right) \mathbf{x}=3 x_{1}^{2}+7 x_{2}^{2}-4 x_{1} x_{2}$
- $Q(\mathbf{x})=\mathbf{x}^{T}\left(\begin{array}{ccc}5 & -\frac{1}{2} & 0 \\ -\frac{1}{2} & 3^{2} & 4 \\ 0 & 4 & 2\end{array}\right) \mathbf{x}=5 x_{1}^{2}+3 x_{2}^{2}+2 x_{3}^{2}-x_{1} x_{2}+8 x_{2} x_{3}$


## Change of variables in quadratic forms

## Change of variables

A change of variables is an equation of the form $\mathbf{x}=P \mathbf{y}$ or equivalently $P^{-1} \mathbf{x}=\mathbf{y}$, where $P$ is an invertible matrix. Exploiting the fact that, in a quadratic form, $A$ is symmetric, then we have $A=P D P^{T}$. We perform the change of variables

$$
\mathbf{x}=P \mathbf{y}
$$

to obtain

$$
Q(\mathbf{x})=(P \mathbf{y})^{T} A(P \mathbf{y})=\mathbf{y}^{T} P^{T} A P \mathbf{y}=Q(\mathbf{y})
$$

But we know

$$
A=P D P^{T} \Rightarrow D=P^{T} A P
$$

Consequently

$$
Q(\mathbf{y})=\mathbf{y}^{T} D \mathbf{y}
$$

That is, there is a basis, in which the matrix of the quadratic form is diagonal.

## Change of variables in quadratic forms

## Example

Consider $Q(\mathbf{x})=\mathbf{x}^{\top} A \mathbf{x}$ with

$$
A=\left(\begin{array}{cc}
1 & -4 \\
-4 & -5
\end{array}\right)=\left(\begin{array}{cc}
\frac{2}{\sqrt{5}} & \frac{1}{\sqrt{5}} \\
-\frac{1}{\sqrt{5}} & \frac{2}{\sqrt{5}}
\end{array}\right)\left(\begin{array}{cc}
3 & 0 \\
0 & -7
\end{array}\right)\left(\begin{array}{cc}
\frac{2}{\sqrt{5}} & -\frac{1}{\sqrt{5}} \\
\frac{1}{\sqrt{5}} & \frac{2}{\sqrt{5}}
\end{array}\right)
$$

That is

$$
Q(\mathbf{x})=x_{1}^{2}-5 x_{2}^{2}-8 x_{1} x_{2}
$$

If we make the change of variable

$$
\mathbf{y}=P^{T} \mathbf{x}=\binom{\frac{2}{\sqrt{5}} x_{1}-\frac{1}{\sqrt{5}} x_{2}}{\frac{1}{\sqrt{5}} x_{1}+\frac{2}{\sqrt{5}} x_{2}}
$$

then

$$
Q(\mathbf{y})=\mathbf{y}^{\top} D \mathbf{y}=3 y_{1}^{2}-7 y_{2}^{2}
$$

## Change of variables in quadratic forms

Let's check that effectively both ways of calculating the quadratic form are equivalent. For doing so, we'll calculate the value of $Q(\mathbf{x})$ for $\mathbf{x}=(2,-2)$ :

$$
Q(\mathbf{x})=\mathbf{x}^{T} A \mathbf{x}=2^{2}-5 \cdot(-2)^{2}-8 \cdot 2 \cdot(-2)=4-20+32=16
$$

If we make the change of variable

$$
\mathbf{y}=\binom{\frac{2}{\sqrt{5}} 2-\frac{1}{\sqrt{5}}(-2)}{\frac{1}{\sqrt{5}} 2+\frac{2}{\sqrt{5}}(-2)}=\binom{\frac{6}{\sqrt{5}}}{-\frac{2}{\sqrt{5}}}
$$

then

$$
Q(\mathbf{y})=\mathbf{y}^{T} D \mathbf{y}=3\left(\frac{6}{\sqrt{5}}\right)^{2}-7\left(-\frac{2}{\sqrt{5}}\right)^{2}=3 \frac{36}{5}-7 \frac{4}{5}=\frac{80}{5}=16
$$

## Change of variables in quadratic forms



## Theorem 2.1 (Principal axes theorem)

Let $A \in \mathcal{M}_{n \times n}$ be a symmetric matrix. Then, there exists a change of variable $\mathbf{x}=P \mathbf{y}$ such that the quadratic form $\mathbf{x}^{\top} A \mathbf{x}$ becomes $\mathbf{y}^{\top} D \mathbf{y}$ with $D$ an $n \times n$ diagonal matrix. The columns of $P$ are the principal axes.

## Principal axes

## A geometric view of the principal axes

Consider the quadratic form $Q(\mathbf{x})=\mathbf{x}^{\top} A \mathbf{x}$ with $\mathbf{x} \in \mathbb{R}^{2}$ and the isocurve $Q(\mathbf{x})=c$. The isocurve is either an ellipse, a circle, a hyperbola, two intersecting lines, a point, or contains no points at all. If $A$ is diagonal, then

$$
Q(\mathbf{x})=a_{11} x_{1}^{2}+a_{22} x_{2}^{2}=c
$$

The equation of an ellipse is

$$
\frac{x_{1}^{2}}{a^{2}}+\frac{x_{2}^{2}}{b^{2}}=1
$$

with $a, b>0$. Therefore

$$
a=\sqrt{\frac{c}{\mathrm{a}_{11}}} \quad b=\sqrt{\frac{c}{\mathrm{a}_{22}}}
$$



$$
\frac{x_{1}^{2}}{a^{2}}+\frac{x_{2}^{2}}{b^{2}}=1, a>b>0
$$

ellipse

## Principal axes

The equation of a hyperbola is

$$
\frac{x_{1}^{2}}{a^{2}}-\frac{x_{2}^{2}}{b^{2}}=1
$$

with $a, b>0$


## Principal axes

If $A$ is not diagonal, then the ellipse or the hyperbola are rotated

(a) $5 x_{1}^{2}-4 x_{1} x_{2}+5 x_{2}^{2}=48$
(b) $x_{1}^{2}-8 x_{1} x_{2}-5 x_{2}^{2}=16$

## Principal axes

## Example

Let's analyze the rotated ellipse

$$
5 x_{1}^{2}-4 x_{1} x_{2}+5 x_{2}^{2}=48
$$

The corresponding matrix is

$$
A=\left(\begin{array}{cc}
5 & -2 \\
-2 & 5
\end{array}\right)=\left(\begin{array}{cc}
\frac{1}{\sqrt{2}} & -\frac{1}{\sqrt{2}} \\
\frac{1}{\sqrt{2}} & \frac{1}{\sqrt{2}}
\end{array}\right)\left(\begin{array}{ll}
3 & 0 \\
0 & 7
\end{array}\right)\left(\begin{array}{cc}
\frac{1}{\sqrt{2}} & \frac{1}{\sqrt{2}} \\
-\frac{1}{\sqrt{2}} & \frac{1}{\sqrt{2}}
\end{array}\right)
$$

So,

$$
a=\sqrt{\frac{c}{a_{11}}}=\sqrt{\frac{48}{3}}=3 \quad b=\sqrt{\frac{c}{a_{22}}}=\sqrt{\frac{48}{7}} \approx 2.65
$$

The change of variable $\mathbf{x}=\left(\begin{array}{cc}\frac{1}{\sqrt{2}} & -\frac{1}{\sqrt{2}} \\ \frac{1}{\sqrt{2}} & \frac{1}{\sqrt{2}}\end{array}\right) \mathbf{y}$ diagonalizes the quadratic form (see the new axes in the previous slide).

## Classification of quadratic forms

## Example

Look at the following surfaces defined as $z=Q(\mathbf{x})$


The curves seen in $\mathbb{R}^{2}$ are the cut of these surfaces with the plane $z=c$. It is obvious that some of the surfaces are always above $z=0$ ( $a$ and $b$ ), others are always below $z=0(\mathrm{~d})$, and still other are sometimes below and sometimes above $z=0$ (c).

## Classification of quadratic forms

## Definition 2.2 (Classification of quadratic forms)

We say $Q(\mathbf{x})$ is

- positive definite if $Q(\mathbf{x})>0 \quad \forall \mathbf{x} \in \mathbb{R}^{n}, \mathbf{x} \neq \mathbf{0}$

- negative definite if $Q(\mathbf{x})<0 \quad \forall \mathbf{x} \in \mathbb{R}^{n}, \mathbf{x} \neq \mathbf{0}$
- indefinite if $Q(\mathbf{x})$ assumes both positive and negative values
- positive semidefinite if $Q(\mathbf{x}) \geq 0 \quad \forall \mathbf{x} \in \mathbb{R}^{n}, \mathbf{x} \neq \mathbf{0}$
- negative semidefinite if $Q(\mathbf{x}) \leq 0 \quad \forall \mathbf{x} \in \mathbb{R}^{n}, \mathbf{x} \neq \mathbf{0}$


Negative definite


Indefinite

## Classification of quadratic forms

Theorem 2.2 (Classification of quadratic forms and quadratic forms)
Let $Q(\mathbf{x})=\mathbf{x}^{\top} A \mathbf{x}$ with $A \in \mathcal{M}_{n \times n}$ and symmetric. Let $\lambda_{i}$ be the eigenvalues of $A$. $Q(\mathbf{x})$ is

- positive definite iff $\lambda_{i}>0 \quad \forall i$
- negative definite iff $\lambda_{i}<0 \quad \forall i$
- indefinite iff there are positive and negative eigenvalues
- positive semidefinite iff $\lambda_{i} \geq 0 \quad \forall i$
- negative semidefinite iff $\lambda_{i} \leq 0 \quad \forall i$


## Proof

By the Theorem of Principal Axes (Theorem 2.1), there is a change of variable such that

$$
Q(\mathbf{y})=\mathbf{y}^{\top} D \mathbf{y}=\lambda_{1} y_{1}^{2}+\lambda_{2} y_{2}^{2}+\ldots+\lambda_{n} y_{n}^{2}
$$

where $\lambda_{i}$ is the $i$-th eigenvalue. The values of $Q$ depend on $\lambda_{i}$ in the way that the theorem states (e.g., $\forall \mathbf{y} \neq \mathbf{0} \quad Q(\mathbf{y})>0$ iff $\lambda_{i}>0 \quad \forall i$, etc.)

## Classification of quadratic forms

## Examples

- $Q(\mathbf{x})=3 x_{1}^{2}+7 x_{2}^{2}$ is positive definite because its eigenvalues are 3 and 7 (both larger than 0 ).
- $Q(\mathbf{x})=3 x_{1}^{2}$ is positive semidefinite because its eigenvalues are 3 and 0 (both larger or equal than 0 ).
- $Q(\mathbf{x})=3 x_{1}^{2}-7 x_{2}^{2}$ is indefinite because its eigenvalues are 3 and -7 (one positive and another negative).
- $Q(x)=-3 x_{1}^{2}-7 x_{2}^{2}$ is negative definite because its eigenvalues are -3 and -7 (both smaller than 0 ).


## Definition 2.3 (Classification of symmetric matrices)

A symmetric matrix is positive definite if its corresponding quadratic form is positive definite. Analogously for the rest of the classification.

## Classification of quadratic forms

## Cholesky factorization

Cholesky factorization factorizes a symmetric matrix $A$ as

$$
A=R^{T} R
$$

being $R$ an upper triangular matrix. $A$ is positive definite if all entries in the diagonal of $R$ are positive.

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 7, Section 2:

- 7.2.1
- 7.2.3
- 7.2 .5
- 7.2.7
- 7.2.19
- 7.2.23
- 7.2.24
- 7.2.26
- 7.2.27


## Outline

(8) Symmetric matrices and quadratic forms

- Diagonalization of symmetric matrices (a)
- Quadratic forms (b)
- Constrained optimization (b)
- Singular Value Decomposition (SVD) (c)


## Constrained optimization

## Introduction

Many problems in engineering or physics are of the form

$$
\begin{array}{cccc}
\min & Q(\mathbf{x}) & & \max
\end{array} \quad Q(\mathbf{x})
$$

## Example

Calculate the minimum and maximum of $Q(\mathbf{x})=9 x_{1}^{2}+4 x_{2}^{2}+3 x_{3}^{2}$ subject to $\|x\|^{2}=1$. Solution
By taking the minimum and maximum coefficient in $Q(\mathbf{x})$ we have

$$
\begin{aligned}
3 x_{1}^{2}+3 x_{2}^{2}+3 x_{3}^{2} & \leq Q(\mathbf{x}) \leq 9 x_{1}^{2}+9 x_{2}^{2}+9 x_{3}^{2} \\
3\left(x_{1}^{2}+x_{2}^{2}+x_{3}^{2}\right) & \leq Q(\mathbf{x}) \leq 9\left(x_{1}^{2}+x_{2}^{2}+x_{3}^{2}\right) \\
3 & \leq Q(\mathbf{x}) \leq 9
\end{aligned}
$$

The minimum value $Q(\mathbf{x})=3$ is attained for $\mathbf{x}=(0,0,1)$, while the maximum value $Q(\mathbf{x})=9$ is attained for $\mathbf{x}=(1,0,0)$. In fact the minimum and maximum values that the constrained quadratic form can take are $\lambda_{\min }$ and $\lambda_{\text {max }}$.

## Constrained optimization

## Example

Calculate the minimum and maximum of $Q(\mathbf{x})=3 x_{1}^{2}+7 x_{2}^{2}$ subject to $\|\mathbf{x}\|^{2}=1$. Solution
$\|\mathbf{x}\|^{2}=1$ is a cylinder in $\mathbb{R}^{3}$ while $z=Q(\mathbf{x})$ is a parabolic surface. The minimum and maximum of the constrained problem are attained among those points belonging to the curve that is the intersection of both surfaces.


FIGURE $1 z=3 x_{1}^{2}+7 x_{2}^{2}$.


FIGURE 2 The intersection of $z=$ $3 x_{1}^{2}+7 x_{2}^{2}$ and the cylinder $x_{1}^{2}+x_{2}^{2}=1$.

## Constrained optimization

## Theorem 3.1

Let $A$ be a symmetric matrix and let

$$
\begin{aligned}
m & =\min \left\{\mathbf{x}^{T} A \mathbf{x} \mid\|\mathbf{x}\|^{2}=1\right\} \\
M & =\max \left\{\mathbf{x}^{T} A \mathbf{x} \mid\|\mathbf{x}\|^{2}=1\right\}
\end{aligned}
$$

Then, $M=\lambda_{\text {max }}$ and $m=\lambda_{\text {min }} . M$ is attained for $\mathbf{x}=\mathbf{u}_{\text {max }}$ (the eigenvector associated to $\lambda_{\text {max }}$ ) and $m$ is attained for $\mathbf{x}=\mathbf{u}_{\text {min }}$ (the eigenvector associated to $\lambda_{\text {min }}$ ).
Proof
Let's orthogonally diagonalize $A$ as $A=P D P^{T}$ and we make the change variables $\mathbf{y}=P^{T} \mathbf{x}$. We already know that

$$
Q(\mathbf{x})=\mathbf{x}^{T} A \mathbf{x}=\mathbf{y}^{T} D \mathbf{y}
$$

Additionally $\|\mathbf{y}\|^{2}=\|\mathbf{x}\|^{2}$ because

$$
\|\mathbf{y}\|^{2}=\mathbf{y}^{T} \mathbf{y}=\left(P^{T} \mathbf{x}\right)^{T}\left(P^{T} \mathbf{x}\right)=\mathbf{x}^{T} P P^{T} \mathbf{x}=\mathbf{x}^{T} \mathbf{x}=\|\mathbf{x}\|^{2}
$$

In particular $\|\mathbf{y}\|=1 \Leftrightarrow\|\mathbf{x}\|=1$.

## Constrained optimization

Then,

$$
\begin{aligned}
m & =\min \left\{\mathbf{y}^{\top} D \mathbf{y} \mid\|\mathbf{y}\|^{2}=1\right\} \\
M & =\max \left\{\mathbf{y}^{\top} D \mathbf{y} \mid\|\mathbf{y}\|^{2}=1\right\}
\end{aligned}
$$

Since $D$ is diagonal we have

$$
\mathbf{y}^{\top} D \mathbf{y}=\lambda_{1} y_{1}^{2}+\lambda_{2} y_{2}^{2}+\ldots+\lambda_{n} y_{n}^{2}
$$

Let's look for the maximum of these values subject to $\|\boldsymbol{y}\|=1$. Consider the maximum eigenvalue, $\lambda_{\max }$, then

$$
\begin{aligned}
\mathbf{y}^{\top} D \mathbf{y} & =\lambda_{1} y_{1}^{2}+\lambda_{2} y_{2}^{2}+\ldots+\lambda_{n} y_{n}^{2} \\
& \leq \lambda_{\max } y_{1}^{2}+\lambda_{\max } y_{2}^{2}+\ldots+\lambda_{\max }^{2} y_{n}^{2} \\
& =\lambda_{\max }\left(y_{1}^{2}+y_{2}^{2}+\ldots+y_{n}^{2}\right) \\
& =\lambda_{\max }\|\mathbf{y}\|=\lambda_{\max }
\end{aligned}
$$

## Constrained optimization

In fact the value $\lambda_{\max }$ is attained for $\mathbf{y}_{\max }=\left(\begin{array}{llllllll}0 & 0 & \ldots & 0 & 1 & 0 & \ldots & 0\end{array}\right)$, where the 1 is at the location corresponding to $\lambda_{\max }$. The corresponding $\mathbf{x}$ is

$$
\mathbf{x}=P \mathbf{y}=\left(\begin{array}{lllllllll}
\mathbf{u}_{1} & \mathbf{u}_{2} & \ldots & \mathbf{u}_{\max -1} & \mathbf{u}_{\max } & \mathbf{u}_{\max +1} & \ldots & \mathbf{u}_{n}
\end{array}\right)\left(\begin{array}{c}
0 \\
0 \\
\ldots \\
0 \\
1 \\
0 \\
\ldots \\
0
\end{array}\right)=\mathbf{u}_{\max }
$$

We could reason analogously for the minimum.

## Constrained optimization

## Example

Let $A=\left(\begin{array}{lll}3 & 2 & 1 \\ 2 & 3 & 1 \\ 1 & 1 & 4\end{array}\right)$. Solve the following optimization problem

$$
\begin{array}{cc}
\max & Q(\mathbf{x})=\mathbf{x}^{\top} A \mathbf{x} \\
\text { subject to } & \|\mathbf{x}\|^{2}=1
\end{array}
$$

## Solution

The characteristic equation is

$$
|A-\lambda I|=0=-(\lambda-6)(\lambda-3)(\lambda-1)
$$

The maximum eigenvalue is $\lambda=6$ and its corresponding eigenvector is $\mathbf{u}=\left(\frac{1}{\sqrt{3}}, \frac{1}{\sqrt{3}}, \frac{1}{\sqrt{3}}\right)$. Therefore, the maximum of $Q(\mathbf{x})$ is 6 that is attained for $\mathbf{x}=\left(\frac{1}{\sqrt{3}}, \frac{1}{\sqrt{3}}, \frac{1}{\sqrt{3}}\right)$.

## Constrained optimization

## Theorem 3.2

Let $A, \lambda_{\max }$ and $\mathbf{u}_{\max }$ be defined as in the previous theorem. Then the solution of

$$
\begin{array}{cc}
\max & Q(\mathbf{x})=\mathbf{x}^{\top} A \mathbf{x} \\
\text { subject to } & \|\mathbf{x}\|^{2}=1 \\
& \mathbf{x} \cdot \mathbf{u}_{\max }=0
\end{array}
$$

is given by the second largest eigenvalue $\lambda_{\max -1}$ that is attained for its associated eigenvector ( $\mathbf{u}_{\text {max }-1}$ ).

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 7, Section 3:

- 7.3.1
- 7.3.3
- 7.3.13


## Outline

(8) Symmetric matrices and quadratic forms

- Diagonalization of symmetric matrices (a)
- Quadratic forms (b)
- Constrained optimization (b)
- Singular Value Decomposition (SVD) (c)


## Singular Value Decomposition (SVD)

## Introduction

Unfortunately, not all matrices can be diagonalized and factorized as

$$
A=P D P^{-1}
$$

However, all of them (even rectangular matrices) can be factorized as

$$
A=Q D P^{-1}
$$

This is called the Singular Value Decomposition. It imitates the property of stretching/shrinking of eigenvalues and eigenvectors. For instance, assume $\mathbf{u}$ is an eigenvector, then

$$
A \mathbf{u}=\lambda \mathbf{u} \Rightarrow\|A \mathbf{u}\|=|\lambda|\|\mathbf{u}\|
$$

If $|\lambda|>1$, then the transformed vector $A \mathbf{u}$ is stretched with respect to $\mathbf{u}$. On the contrary, if $|\lambda|<1$, then the transformed vector $A \mathbf{u}$ is shrinked with respect to $\mathbf{u}$.

## Singular Value Decomposition (SVD)

## Example

Consider $A=\left(\begin{array}{ccc}4 & 11 & 14 \\ 8 & 7 & -2\end{array}\right)$ and the linear transformation $T(\mathbf{x})=A \mathbf{x}$. It transforms the unit sphere in $\mathbb{R}^{3}$ onto an ellipse of $\mathbb{R}^{2}$


FIGURE 1 A transformation from $\mathbb{R}^{3}$ to $\mathbb{R}^{2}$.
Look for the direction that maximizes $\|A \mathbf{x}\|$ subject to $\|\mathbf{x}\|=1$.

## Singular Value Decomposition (SVD)

## Solution

We may maximize $\|A x\|^{2}$ because $\|A x\|$ is maximum iff $\|A x\|^{2}$ is maximum.

$$
\|A \mathbf{x}\|^{2}=(A \mathbf{x})^{T}(A \mathbf{x})=\mathbf{x}^{T} A^{T} A \mathbf{x}
$$

which is a quadratic form since $A^{T} A$ is symmetric:

$$
A^{T} A=\left(\begin{array}{ccc}
80 & 100 & 40 \\
100 & 170 & 140 \\
40 & 140 & 200
\end{array}\right)
$$

By Theorem 3.1, the maximum eigenvalue is $\max \|A x\|^{2}=\lambda_{\max }=360$ and its associated eigenvector $\mathbf{u}_{\max }=\left(\frac{1}{3}, \frac{2}{3}, \frac{2}{3}\right)$. Consequently $\max \|A \mathbf{x}\|=\sqrt{360}=6 \sqrt{10}$ that is attained for

$$
A \mathbf{u}_{\max }=\binom{18}{6}
$$

## Singular Value Decomposition (SVD)

## Definition 4.1

Singular Values of a matrix Let $A \in \mathcal{M}_{m \times n}$. $A^{T} A$ can always be orthogonally diagonalized. Let $\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{n}\right\}$ a base of $\mathbb{R}^{n}$ formed by the eigenvectors of $A^{T} A$ and let $\lambda_{1}, \lambda_{2}, \ldots, \lambda_{n}$ be its corresponding eigenvalues. Then

$$
\left\|A \mathbf{v}_{i}\right\|^{2}=\left(A \mathbf{v}_{i}\right)^{T}\left(A \mathbf{v}_{i}\right)=\mathbf{v}_{i}^{T} A^{T} A \mathbf{v}_{i}=\mathbf{v}_{i}^{T}\left(\lambda_{i} \mathbf{v}_{i}\right)=\lambda_{i}\left\|\mathbf{v}_{i}\right\|^{2}
$$

If we take the square root

$$
\left\|A \mathbf{v}_{i}\right\|=\sqrt{\lambda_{i}}\left\|\mathbf{v}_{i}\right\|
$$

That is, $\sqrt{\lambda_{i}}$ reflects the amount by which $\mathbf{v}_{i}$ is stretched or shrinked. $\sqrt{\lambda_{i}}$ is called a singular value and it is denoted as $\sigma_{i}$.

## Singular Value Decomposition (SVD)

## Example (continued)



FIGURE 1 A transformation from $\mathbb{R}^{3}$ to $\mathbb{R}^{2}$.

In the example of Slide 45, the singular values are the lengths of the ellipse in $\mathbb{R}^{2}$ and they are $6 \sqrt{10}, 3 \sqrt{10}$ and 0 . From the singular values we learn that the unit sphere in $\mathbb{R}^{3}$ (there are 3 singular values) is collapsed in 2D (one of the singular values is 0 ) onto an ellipse (the remaining two singular values are different from each other).

## Singular Value Decomposition (SVD)

## Theorem 4.1

Let $\left\{\mathbf{v}_{1}, \mathbf{v}_{2}, \ldots, \mathbf{v}_{n}\right\}$ a basis of $\mathbb{R}^{n}$ formed by the eigenvectors of $A^{T} A$ sorted in descending order and let $\lambda_{1}, \lambda_{2}, \ldots, \lambda_{n}$ be its corresponding eigenvalues. Let us assume that $A$ has $r$ non-null singular values. Then

$$
S=\left\{A \mathbf{v}_{1}, A \mathbf{v}_{2}, \ldots, A \mathbf{v}_{r}\right\}
$$

is a basis of $\operatorname{Col}\{A\}$ and

$$
\operatorname{Rank}\{A\}=r
$$

## Proof

By Theorem 1.1, any two eigenvectors are orthogonal to each other if they correspond to different eigenvalues, that is, $\mathbf{v}_{i} \cdot \mathbf{v}_{j}=0$. Then,

$$
\left(A \mathbf{v}_{i}\right) \cdot\left(A \mathbf{v}_{j}\right)=\mathbf{v}_{i}^{T} A^{T} A \mathbf{v}_{j}=\mathbf{v}_{i}^{T}\left(\lambda_{j} \mathbf{v}_{j}\right)=\lambda_{j}\left(\mathbf{v}_{i}^{T} \mathbf{v}_{j}\right)=\lambda_{j}\left(\mathbf{v}_{i} \cdot \mathbf{v}_{j}\right)=0
$$

That is $A \mathbf{v}_{i}$ and $A \mathbf{v}_{j}$ are also orthogonal.

## Singular Value Decomposition (SVD)

Additionally, if the eigenvectors $\mathbf{v}_{i}$ are unitary, then (see Definition 4.1)

$$
\sigma_{i}=\left\|A \mathbf{v}_{i}\right\|
$$

Since there are $r$ non-null singular values, $A \mathbf{v}_{i} \neq \mathbf{0}$ only for $i=1,2, \ldots, r$. So the set $S$ is a set of non-null, orthogonal vectors. To show it is a basis of $\operatorname{Col}\{A\}$ we still need to show that any vector in $\operatorname{Col}\{A\}$ can be expressed as a linear combination of the vectors in $S$. We know that the eigenvalues of $A^{T} A$ is a basis of $\mathbb{R}^{n}$. Then for any vector $\mathbf{x} \in \mathbb{R}^{n}$ there exist coefficients $c_{1}, c_{2}, \ldots, c_{n}$ not all of them zero such that

$$
\mathbf{x}=c_{1} \mathbf{v}_{1}+c_{2} \mathbf{v}_{2}+\ldots+c_{n} \mathbf{v}_{n}
$$

If we transform this vector

$$
\begin{aligned}
A \mathbf{x} & =A\left(c_{1} \mathbf{v}_{1}+c_{2} \mathbf{v}_{2}+\ldots+c_{n} \mathbf{v}_{n}\right) & & \text { [Linear transformation] } \\
& =c_{1} A \mathbf{v}_{1}+c_{2} A \mathbf{v}_{2}+\ldots+c_{n} A \mathbf{v}_{n} & & \text { [non-null singular values] } \\
& =c_{1} A \mathbf{v}_{1}+c_{2} A \mathbf{v}_{2}+\ldots+c_{r} A \mathbf{v}_{r} & &
\end{aligned}
$$

## Singular Value Decomposition (SVD)

That is any transformed vector $A \mathbf{x}$ can be expressed as a linear combination of the elements in $S$. Consequently, $S$ is a basis of $\operatorname{Col}\{A\}$. Finally, $\operatorname{Rank}\{A\}$ is nothing more than the dimension of $\operatorname{Col}\{A\}$. Since $A$ is a basis of $\operatorname{Col}\{A\}$ and it has $r$ vectors, then $\operatorname{Rank}\{A\}=r$.

## Singular Value Decomposition (SVD)

## Theorem 4.2 (The Singular Value Decomposition)

Let $A \in \mathcal{M}_{m \times n}$ be a matrix with rank $r$. Then, there exists a matrix $\Sigma \in \mathcal{M}_{m \times n}$ whose diagonal entries are the first $r$ singular values of $A$ sorted in descending order ( $\sigma_{1} \geq \sigma_{2} \geq \ldots \geq \sigma_{r}>0$ ) and there exist orthogonal matrices $U \in \mathcal{M}_{m \times m}$ and $V \in \mathcal{M}_{n \times n}$ such that

$$
A=U \Sigma V^{T}
$$

$\Sigma$ is unique but $U$ and $V$ are not. The columns of $U$ are called the left singular vectors, and the columns of $V$ are the right singular vectors.

## Example

$$
\left(\begin{array}{llll}
a_{11} & a_{12} & a_{13} & a_{14} \\
a_{21} & a_{22} & a_{23} & a_{24}
\end{array}\right)=\left(\begin{array}{ll}
u_{11} & u_{12} \\
u_{21} & u_{22}
\end{array}\right)\left(\begin{array}{cccc}
\sigma_{1} & 0 & 0 & 0 \\
0 & \sigma_{2} & 0 & 0
\end{array}\right)\left(\begin{array}{llll}
v_{11} & v_{21} & v_{31} & v_{41} \\
v_{12} & v_{22} & v_{32} & v_{42} \\
v_{13} & v_{23} & v_{33} & v_{43} \\
v_{14} & v_{24} & v_{34} & v_{44}
\end{array}\right)
$$

## Singular Value Decomposition (SVD)

## Proof

Let $\lambda_{i}$ and $\mathbf{v}_{i}(i=1,2, \ldots, n)$ be the eigenvalues and eigenvectors of $A^{T} A$. By Theorem 4.1 we know that $S=\left\{A \mathbf{v}_{1}, A \mathbf{v}_{2}, \ldots, A \mathbf{v}_{r}\right\}$ is an orthogonal basis of $\operatorname{Col}\{A\}$. Let's normalize these vectors

$$
\mathbf{u}_{i}=\frac{A \mathbf{v}_{i}}{\sigma_{i}} \quad i=1,2, \ldots, r
$$

and we extend the set $\left\{\mathbf{u}_{1}, \mathbf{u}_{2}, \ldots, \mathbf{u}_{r}\right\}$ to be an orthogonal basis of $\mathbb{R}^{m}$. Let us construct the matrices

$$
\begin{aligned}
& U=\left(\begin{array}{llll}
\mathbf{u}_{1} & \mathbf{u}_{2} & \ldots & \mathbf{u}_{m}
\end{array}\right) \\
& V=\left(\begin{array}{llll}
\mathbf{v}_{1} & \mathbf{v}_{2} & \ldots & \mathbf{v}_{n}
\end{array}\right)
\end{aligned}
$$

By construction $U$ and $V$ are orthogonal, and

$$
\begin{aligned}
A V & =\left(\begin{array}{lllllll}
A \mathbf{v}_{1} & A \mathbf{v}_{2} & \ldots & A \mathbf{v}_{r} & \mathbf{0} & \ldots & \mathbf{0}
\end{array}\right) \\
& =\left(\begin{array}{llllll}
\sigma_{1} \mathbf{u}_{1} & \sigma_{2} \mathbf{u}_{2} & \ldots & \sigma_{r} \mathbf{u}_{r} & \mathbf{0} & \ldots
\end{array}\right)
\end{aligned}
$$

## Singular Value Decomposition (SVD)

Proof (continued)
On the other side, let

$$
D=\left(\begin{array}{cccc}
\sigma_{1} & 0 & \ldots & 0 \\
0 & \sigma_{2} & \ldots & 0 \\
\ldots & \ldots & \ldots & \ldots \\
0 & 0 & \ldots & \sigma_{r}
\end{array}\right) \quad \Sigma=\left(\begin{array}{ll}
D & 0 \\
0 & 0
\end{array}\right)
$$

Then,

$$
U \Sigma=\left(\begin{array}{llll}
\mathbf{u}_{1} & \mathbf{u}_{2} & \ldots & \mathbf{u}_{m}
\end{array}\right)\left(\begin{array}{ll}
D & 0 \\
0 & 0
\end{array}\right)=\left(\begin{array}{lllllll}
\sigma_{1} \mathbf{u}_{1} & \sigma_{2} \mathbf{u}_{2} & \ldots & \sigma_{r} \mathbf{u}_{r} & \mathbf{0} & \ldots & \mathbf{0}
\end{array}\right)
$$

Therefore,

$$
U \Sigma=A V \Rightarrow A=U \Sigma V^{\top}
$$

since $V$ is orthogonal.

## Singular Value Decomposition (SVD)

## Theorem 4.3 (Properties of the SVD decomposition)

In a SVD decomposition

- The left singular vectors of $A$ are eigenvectors of $A A^{T}$.
- The right singular vectors of $A$ are eigenvectors of $A^{T} A$.
- The singular values are the square root of the eigenvalues of both $A A^{T}$ and $A^{T} A$.
- The singular values are the length of the semiaxes of the mapping of the unit hypersphere in $\mathbb{R}^{n}$ onto $\mathbb{R}^{m}$.
- The columns of $U$ form an orthogonal basis of $\mathbb{R}^{m}$.
- The columns of $V$ form an orthogonal basis of $\mathbb{R}^{n}$.


## Singular Value Decomposition (SVD)

## Example

Let's calculate the SVD decomposition of $A=\left(\begin{array}{ccc}4 & 11 & 14 \\ 8 & 7 & -2\end{array}\right)$.
Step 1: Orthogonally diagonalize $A^{T} A$

$$
A^{T} A=\left(\begin{array}{ccc}
80 & 100 & 40 \\
100 & 170 & 140 \\
40 & 140 & 200
\end{array}\right)
$$

Its eigenvalues and eigenvectors are

$$
\begin{aligned}
\lambda_{1}=360 & \mathbf{v}_{1}=\left(\frac{1}{3}, \frac{2}{3}, \frac{2}{3}\right) \\
\lambda_{2}=90 & \mathbf{v}_{2}=\left(-\frac{2}{3},-\frac{1}{3}, \frac{2}{3}\right) \\
\lambda_{3}=0 & \mathbf{v}_{3}=\left(\frac{2}{3},-\frac{2}{3}, \frac{1}{3}\right)
\end{aligned}
$$

## Singular Value Decomposition (SVD)

Step 2: Construct $V$ and $\Sigma$

$$
\begin{aligned}
V=\left(\begin{array}{lll}
\mathbf{v}_{1} & \mathbf{v}_{2} & \mathbf{v}_{3}
\end{array}\right)=\left(\begin{array}{ccc}
\frac{1}{3} & -\frac{2}{3} & \frac{2}{3} \\
\frac{2}{3} & -\frac{1}{3} & -\frac{2}{3} \\
\frac{2}{3} & \frac{2}{3} & \frac{1}{3}
\end{array}\right) \\
\Sigma=\left(\begin{array}{ccc}
\sqrt{\lambda_{1}} & 0 & 0 \\
0 & \sqrt{\lambda_{2}} & 0
\end{array}\right)=\left(\begin{array}{ccc}
6 \sqrt{10} & 0 & 0 \\
0 & 3 \sqrt{10} & 0
\end{array}\right)
\end{aligned}
$$

Step 3: Construct $U$

$$
\begin{gathered}
\mathbf{u}_{1}=\frac{A \mathbf{v}_{1}}{\sigma_{1}}=\left(\frac{3}{\sqrt{10}}, \frac{1}{\sqrt{10}}\right) \\
\mathbf{u}_{2}=\frac{A \mathbf{v}_{2}}{\sigma_{2}}=\left(\frac{1}{\sqrt{10}},-\frac{3}{\sqrt{10}}\right)
\end{gathered}
$$

The set $\left\{\mathbf{u}_{1}, \mathbf{u}_{2}\right\}$ is already a basis of $\mathbb{R}^{2}$, so there is no need to extend it.

## Singular Value Decomposition (SVD)

Finally we have

$$
\begin{gathered}
A=U \Sigma V^{T} \\
\left(\begin{array}{ccc}
4 & 11 & 14 \\
8 & 7 & -2
\end{array}\right)=\left(\begin{array}{ccc}
\frac{3}{\sqrt{10}} & \frac{1}{\sqrt{10}} \\
\frac{1}{\sqrt{10}} & -\frac{3}{\sqrt{10}}
\end{array}\right)\left(\begin{array}{cccc}
6 \sqrt{10} & 0 & 0 \\
0 & 3 \sqrt{10} & 0
\end{array}\right)\left(\begin{array}{ccc}
\frac{1}{3} & \frac{2}{3} & \frac{2}{3} \\
-\frac{2}{3} & -\frac{1}{3} & \frac{2}{3} \\
\frac{2}{3} & -\frac{2}{3} & \frac{1}{3}
\end{array}\right)
\end{gathered}
$$

MATLAB: $[\mathrm{U}, \mathrm{S}, \mathrm{V}]=\operatorname{svd}\left(\left[\begin{array}{lllll}4 & 11 & 14 ; & 8 & 7\end{array}\right]\right)$

## Singular Value Decomposition (SVD)

## Example

Let's calculate the SVD decomposition of $A=\left(\begin{array}{cc}1 & -1 \\ -2 & 2 \\ 2 & -2\end{array}\right)$.
Step 1: Orthogonally diagonalize $A^{T} A$

$$
A^{T} A=\left(\begin{array}{cc}
9 & -9 \\
-9 & 9
\end{array}\right)
$$

Its eigenvalues and eigenvectors are

$$
\begin{aligned}
\lambda_{1}=18 & \mathbf{v}_{1}=\left(\frac{1}{\sqrt{2}},-\frac{1}{\sqrt{2}}\right) \\
\lambda_{2}=0 & \mathbf{v}_{2}=\left(\frac{1}{\sqrt{2}}, \frac{1}{\sqrt{2}}\right)
\end{aligned}
$$

## Singular Value Decomposition (SVD)

Step 2: Construct $V$ and $\Sigma$

$$
\begin{gathered}
V=\left(\begin{array}{ll}
\mathbf{v}_{1} & \mathbf{v}_{2}
\end{array}\right)=\left(\begin{array}{cc}
\frac{1}{\sqrt{2}} & \frac{1}{\sqrt{2}} \\
-\frac{1}{\sqrt{2}} & \frac{1}{\sqrt{2}}
\end{array}\right) \\
\Sigma=\left(\begin{array}{cc}
\sqrt{\lambda_{1}} & 0 \\
0 & \sqrt{\lambda_{2}} \\
0 & 0
\end{array}\right)=\left(\begin{array}{cc}
3 \sqrt{2} & 0 \\
0 & 0 \\
0 & 0
\end{array}\right)
\end{gathered}
$$

Step 3: Construct $U$

$$
\mathbf{u}_{1}=\frac{A \mathbf{v}_{1}}{\sigma_{1}}=\left(\frac{1}{3},-\frac{2}{3}, \frac{2}{3}\right)
$$

The set $\left\{\mathbf{u}_{1}\right\}$ is not yet a basis of $\mathbb{R}^{3}$, so we need to extend it with orthogonal vectors. All vectors orthogonal to $\mathbf{u}_{1}$ fulfill

$$
\mathbf{u}_{1} \cdot \mathbf{u}=0=\frac{1}{3} x_{1}-\frac{2}{3} x_{2}+\frac{2}{3} x_{3} \Rightarrow x_{1}=2 x_{2}-2 x_{3}
$$

## Singular Value Decomposition (SVD)

Step 3: Construct $U$ (continued)
 orthogonal. Let's make it orthogonal following Gram-Schmidt procedure

$$
\begin{aligned}
& \mathbf{u}_{2} \frac{\mathbf{w}_{2}}{\left\|\mathbf{w}_{2}\right\|}=\left(\frac{2}{\sqrt{5}}, \frac{1}{\sqrt{5}}, 0\right) \\
& \mathbf{w}_{3}^{\prime}=\mathbf{w}_{3}-<\mathbf{w}_{3}, \mathbf{v}_{2}>\mathbf{v}_{2}=\left(-\frac{2}{5}, \frac{4}{5}, 1\right) \\
& \mathbf{u}_{3}=\frac{\mathbf{w}_{3}}{\| \mathbf{w}_{3}}=\left(-\frac{2}{3 \sqrt{5}}, \frac{4}{3 \sqrt{5}}, \frac{\sqrt{5}}{3}\right)
\end{aligned}
$$

In fact, SVD does not require the $\mathbf{u}$ vectors to be unitary, but it is simply convenient. We can make $\mathbf{u}_{2}$ and $\mathbf{u}_{3}$ unitary because they are "free" (we are constructing them simply to extend the set of $\mathbf{u}$ vectors to be a basis of $\mathbb{R}^{3}$ ), but not $\mathbf{u}_{1}$ because it is "bound" to the singular value.

## Singular Value Decomposition (SVD)

Finally we have

$$
\left(\begin{array}{cc}
1 & -1 \\
-2 & 2 \\
2 & -2
\end{array}\right)=\left(\begin{array}{ccc}
\frac{1}{3} & \frac{2}{\sqrt{5}} A=U \Sigma V^{T} & -\frac{2}{3 \sqrt{5}} \\
-\frac{2}{3} & \frac{1}{\sqrt{5}} & \frac{4}{3 \sqrt{5}} \\
\frac{2}{3} & 0 & \frac{\sqrt{5}}{3}
\end{array}\right)\left(\begin{array}{cc}
3 \sqrt{2} & 0 \\
0 & 0 \\
0 & 0
\end{array}\right)\left(\begin{array}{cc}
\frac{1}{\sqrt{2}} & -\frac{1}{\sqrt{2}} \\
\frac{1}{\sqrt{2}} & \frac{1}{\sqrt{2}}
\end{array}\right)
$$

## Algebraic applications of SVD

## Matrix condition number

Let $\sigma_{1}$ and $\sigma_{r}$ be the largest and smallest singular values of a matrix $A$. The condition number of the matrix is defined as

$$
\kappa(A)=\frac{\sigma_{1}}{\sigma_{r}}
$$

If this condition number is very large, the equations system $A \mathbf{x}=\mathbf{b}$ is ill-posed and small perturbations in $\mathbf{b}$ translate into large perturbations in $\mathbf{x}$. As a rule of thumb, if $\kappa(A)=10^{k}$, then you may lose up to $k$ digits of accuracy.

## Algebraic applications of SVD

## Bases for fundamental spaces

The $U$ and $V$ matrices provide bases for $\operatorname{Row}\{A\}, \operatorname{Col}\{A\}=\operatorname{Row}\left\{A^{T}\right\}, \operatorname{Nul}\{A\}$ and $\operatorname{Nul}\left\{A^{T}\right\}$


FIGURE 4 The four fundamental subspaces and the action of $A$.

## Algebraic applications of SVD

## Theorem 4.4 (The Invertible Matrix Theorem (continued))

The Invertible Matrix Theorem has been developed in Theorems 5.1 and 11.5 of Chapter 3, Theorem 10.5 of Chapter 5, Theorem 2.1 of Chapter 6. Here, we give an extension if $A$ is invertible, then the following statements are equivalent to the previous statements:
$\mathrm{xxvii} .(\operatorname{Col}\{A\})^{\perp}=\{\mathbf{0}\}$.
xxviii. $(\operatorname{Nul}\{A\})^{\perp}=\mathbb{R}^{n}$.
xxix. $(\operatorname{Row}\{A\})=\mathbb{R}^{n}$.
xxx . $A$ has $n$ non-null singular values.

## Algebraic applications of SVD

## Reduced SVD and pseudoinverse of A

If within $U$ and $V$ we distinguish two submatrices, each one with $r$ columns we have

$$
U=\left(U_{r} U_{m-r}\right) \text { and } V=\left(V_{r} V_{n-r}\right)
$$

Then,

$$
A=U \Sigma V^{T}=\left(U_{r} U_{m-r}\right)\left(\begin{array}{cc}
D & 0 \\
0 & 0
\end{array}\right)\binom{V_{r}^{T}}{V_{n-r}^{T}}=U_{r} D V_{r}^{T}
$$

Despite the fact that we may have removed many columns of $U$ and $V$, we have not lost any information and the recovery of $A$ is exact. The Moore-Penrose pseudoinverse is defined as

$$
A^{+}=V_{r} D^{-1} U_{r}^{T}
$$

that is a $n \times m$ matrix such that

$$
A^{+} A A^{+}=A^{+} \quad A A^{+} A=A
$$

## Algebraic applications of SVD

## Pseudoinverse of $A$ and Least Squares

It can be shown that the least-squares solution of the equation system $A \mathbf{x}=\mathbf{b}$ is given by

$$
\hat{\mathbf{x}}=A^{+} \mathbf{b}
$$

## Matrix approximation

If instead of taking $r$ components in the split of $U$ and $V$ (see previous slide) we take only $k$ (assuming singular values have been ordered in descending order), and we reconstruct $A_{k}$

$$
A_{k}=U_{k} D_{k} V_{k}^{T}
$$

This matrix is the matrix of rank $k$ that minimizes the Frobenius norm of the difference

$$
A_{k}=\min _{\operatorname{Rank}\{B\}=k}\|A-B\|_{F}^{2}=\min _{\operatorname{Rank}\{B\}=k} \sum_{i, j=1}^{n}\left(a_{i j}-b_{i j}\right)^{2}
$$

## Exercises

## Exercises

From Lay (3rd ed.), Chapter 7, Section 4:

- 7.4.3
- 7.4.11
- 7.4.15
- 7.4.17
- 7.4.18
- 7.4.19
- 7.4.20
- 7.4.23
- 7.4.24


## Applications of SVD

## Eigengenes and eigenassays

SVD is very much used to analyze the response of different genes to different assays or conditions.


## Applications of SVD

## Eigengenes and eigenassays

SVD is very much used to analyze the response of different genes to different assays or conditions.


Alter, O., Brown, P. O. and Botstein, D. (2000) Proc. Natl. Acad. Sci. USA 97, 10101

## Applications of SVD

## Eigenfaces

In this example we see the effect of matrix approximation by the reduced SVD.


## Applications of SVD

## Eigenfaces

## We can also use SVD to automatically analyze documents.



[^2]
## Outline

(8) Symmetric matrices and quadratic forms

- Diagonalization of symmetric matrices (a)
- Quadratic forms (b)
- Constrained optimization (b)
- Singular Value Decomposition (SVD) (c)


# Chapter 9. Linear algebra applications in geometry 

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## Outline

(9) Linear algebra applications in geometry

- Local and global coordinates
- Points and vectors
- Lines in 2D
- Affine maps in 2D
- Conic sections in 2D
- 3D Geometry
- Quadrics in 3D


## References


G. Farin, D. Hansford. Practical Linear Algebra: a geometry toolbox. A.K. Peters (2005).
J. de Burgos. Álgebra lineal y geometría cartesiana. McGraw Hill 2a Ed. (2000)

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(9) Linear algebra applications in geometry

- Local and global coordinates
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- Quadrics in 3D


## Local and global coordinates

## Reference

Farin and Hansford, Chapter 1

## Local and global coordinates

In real applications we may need to distinguish between local and global coordinates.


And we need some way of transforming one into the other. This is nothing more than a change of basis.

## Local and global coordinates

## Shift and scale

In Vector Graphics it is common to design objects in a local coordinate system (d) and, then, place, rotate and scale the object in the global coordinate system (e). We need some transformation to go from one space to the other.

For the first component, $d_{1}$, we note that we go from a local interval $[0,1]$ to a global interval $\left[\min _{1}\right.$, max $\left._{1}\right]$. We may easily perform the transformation as

$$
\begin{gathered}
\frac{d_{1}-0}{1-0}=\frac{e_{1}-\min _{1}}{\max _{1} \min _{1}} \Rightarrow \\
e_{1}=\min _{1}+\left(\max _{1}-\min _{1}\right) d_{1}
\end{gathered}
$$



## Local and global coordinates

## Shift and scale

The more general transformation maps the local interval $\left[\min _{d 1}, \max _{d 1}\right]$ to the global interval $\left[\min _{e 1}, \max _{e 1}\right]$. This is achieved with transformation

$$
e_{1}=m i n_{e 1}+\frac{m a x_{e 1}-m i_{e 1}}{\max _{d 1}-\min _{d 1}} d_{1}
$$

The same kind of transformation is applied to the second component $\left(d_{2} \rightarrow e_{2}\right)$. Putting everything in matrix notation we have

$$
\mathbf{e}=\binom{\min _{e 1}}{\min _{e 2}}+\left(\begin{array}{cc}
\frac{m a x_{e l}-\min _{e 1}}{\max _{d 1}-\min _{d 1}} & 0 \\
0 & \frac{\text { max }_{e 2}-\min _{e 2}}{\max _{d 2}-\min _{d 2}}
\end{array}\right) \mathbf{d}
$$

This transformation is of the form

$$
\mathbf{e}=T(\mathbf{d})=\mathbf{e}_{\min }+A \mathbf{d}
$$

that is not a linear transformation because of the shift (e.g., show that $\left.T\left(\mathbf{d}_{1}+\mathbf{d}_{2}\right) \neq T\left(\mathbf{d}_{1}\right)+T\left(\mathbf{d}_{2}\right)\right)$.

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## Points and vectors

## Reference

Farin and Hansford, Chapter 2

## Points and vectors

We also need to distinguish between points and vectors. Both are represented as a list of coordinates. Informally, a point indicates a location in space, while a vector indicates a direction (orientation+sense) in space. In this example, we have two points, $\mathbf{p}$ and $\mathbf{q}$, and a vector $\mathbf{v}$ that goes from $\mathbf{p}$ to $\mathbf{q}$. We may talk about the length of a vector, but not of a point.


## Points and vectors

## Points and vectors

In this example we have multiple copies of the same vector (since they all have the same direction and magnitude). In Physics, forces are vectors that are applied to objects that are located at points. In this figure we would see the same force applied to different objects.


## Points and vectors

## Points and vectors

More formally, points belong to an Euclidean space while vectors belong to a vector space.

$$
\begin{gathered}
\mathbf{p}, \mathbf{q} \in \mathbb{E}^{2} \\
\mathbf{v} \in \mathbb{R}^{2}
\end{gathered}
$$

Although we may represent both spaces in the same figure and we may define operations using both kinds of spaces. The goal of distinguishing between points and vectors is to distinguish between operations that depend on the coordinate system and operations that do not.


## Operations on points and vectors

Coordinate independent operations
$-: \mathbb{E}^{2} \times \mathbb{E}^{2} \rightarrow \mathbb{R}^{2} \quad \mathbf{v}=\mathbf{q}-\mathbf{p}$
$+: \mathbb{E}^{2} \times \mathbb{R}^{2} \rightarrow \mathbb{E}^{2} \quad \mathbf{p}=\mathbf{q}+\mathbf{v}$
$+: \mathbb{R}^{2} \times \mathbb{R}^{2} \rightarrow \mathbb{R}^{2} \quad \mathbf{v}=\mathbf{u}+\mathbf{w}$
$\cdot: \quad \mathbb{R} \times \mathbb{R}^{2} \rightarrow \mathbb{R}^{2} \quad \mathbf{v}=r \mathbf{u}$

Coordinate dependent operations

$$
\begin{array}{rll}
+: & \mathbb{E}^{2} \times \mathbb{E}^{2} \rightarrow \mathbb{E}^{2} & \mathbf{t}=\mathbf{p}+\mathbf{q} \\
\therefore & \mathbb{R} \times \mathbb{E}^{2} \rightarrow \mathbb{E}^{2} & \mathbf{q}=r \mathbf{p}
\end{array}
$$

## Vector fields

## Vector fields

Any function that assigns a vector to a point $f: \mathbb{E}^{2} \rightarrow \mathbb{R}^{2} \quad \mathbf{v}=f(\mathbf{p})$

## Example

$$
f(x, y)=(x, y)
$$


$f(x, y)=(-y, x)$


## Combinations of points

## Barycentric combinations

A weighted sum of points where the weights add up to 1 is called a barycentric combination

## Example

$$
\mathbf{r}=(1-t) \mathbf{p}+t \mathbf{q}=\mathbf{p}+t(\mathbf{q}-\mathbf{p})
$$

$$
\mathbf{s}=t_{1} \mathbf{r}+t_{2} \mathbf{p}+t_{3} \mathbf{q}
$$



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## Lines in 2D

## Reference

Farin and Hansford, Chapter 3
Parametric equation of a line

- Given two points:
- Given point and vector: $\overline{\mathbf{I}}(t)=\mathbf{p}+t \mathbf{v} \quad t \in \mathbb{R}$



## Lines in 2D

Implicit equation of a line

- Given a point and the normal direction: $\mathbf{a} \cdot(\mathbf{x}-\mathbf{p})=0$

In 2D:

$$
\begin{gathered}
\left(a_{1}, a_{2}\right) \cdot\left(x_{1}-p_{1}, x_{2}-p_{2}\right)=0 \Rightarrow \\
a x_{1}+b x_{2}+c=0
\end{gathered}
$$



## Lines in 2D

## Explicit equation of a line

- Given a point and slope: In 2D:

$$
\begin{gathered}
x_{2}=p_{2}+m\left(x_{1}-p_{1}\right) \\
x_{2}=m x_{1}+b \\
x_{2}=(\tan \Theta) x_{1}+b
\end{gathered}
$$

But it is not a good representation for vertical lines.


## Lines in 2D

## Distance of a point to a line

- Implicit line:

Line: $\mathbf{a} \cdot(\mathbf{x}-\mathbf{p})=0$
Point: $\mathbf{r}$

Let $\mathbf{w}=\mathbf{r}-\mathbf{p}$ and calculate:

$$
\mathbf{a} \cdot \mathbf{w}=\|\mathbf{a}\|\|\mathbf{w}\| \cos (\theta)
$$

Analyzing the figure we note that $\cos (\theta)=\frac{d}{\|w\|}$. Then

$$
\mathbf{a} \cdot \mathbf{w}=\|\mathbf{a}\| d \Rightarrow d=\frac{\mathbf{a} \cdot \mathbf{w}}{\|\mathbf{a}\|}
$$



## Lines in 2D

## Distance of a point to a line

- Parametric line:

Line: $\mathbf{I}(t)=\mathbf{p}+t \mathbf{v}$
Point: $\mathbf{r}$

Let $\mathbf{w}=\mathbf{r}-\mathbf{p}$ and calculate:

$$
\mathbf{v} \cdot \mathbf{w}=\|\mathbf{v}\|\|\mathbf{w}\| \cos (\alpha)
$$

Analyzing the figure we note that $\sin (\alpha)=\frac{d}{\|w\|}=\sqrt{1-\cos ^{2}(\alpha)}$. Then

$$
d=\|\mathbf{w}\| \sqrt{1-\left(\frac{\mathbf{v} \cdot \mathbf{w}}{\|\mathbf{v}\|\|\mathbf{w}\|}\right)^{2}}
$$



## Lines in 2D

The foot of a point

- Parametric line:

Line: $\mathbf{I}(t)=\mathbf{p}+t \mathbf{v}$
Point: $\mathbf{r}$

Let $\mathbf{w}=\mathbf{r}-\mathbf{p}$. The closest point within the line to $\mathbf{r}$ is

$$
\mathbf{q}=\mathbf{p}+\operatorname{Proj}_{\mathbf{v}}\{\mathbf{w}\}=\mathbf{p}+\frac{\mathbf{v} \cdot \mathbf{w}}{\|\mathbf{v}\|^{2} \mathbf{v}}
$$



## Lines in 2D

## The intersection of two lines

- Parametric lines:

Line 1: $\mathbf{I}_{1}(t)=\mathbf{p}+t \mathbf{v}$
Line 2: $\mathbf{I}_{2}(s)=\mathbf{q}+s \mathbf{w}$

We need to solve the equation system

$$
\begin{gathered}
\mathbf{l}_{1}(t)=\mathbf{I}_{2}(s) \\
\mathbf{p}+t \mathbf{v}=\mathbf{q}+s \mathbf{w} \\
\left(\begin{array}{ll}
\mathbf{v} & -\mathbf{w}
\end{array}\right)\binom{t}{s}=\mathbf{q}-\mathbf{p}
\end{gathered}
$$



## Lines in 2D

## The intersection of two lines

- Implicit lines:

Line 1: $\mathbf{a} \cdot(\mathbf{x}-\mathbf{p})=0$
Line 1: $\overline{\mathbf{a}} \cdot(\mathbf{x}-\mathbf{q})=0$
We need to find x satisfying both equations at the same time

$$
\begin{array}{r}
\mathbf{a}^{T} \mathbf{x}-\mathbf{a}^{T} \mathbf{p}=0 \\
\overline{\mathbf{a}}^{T} \mathbf{x}-\overline{\mathbf{a}}^{T} \mathbf{q}=0 \\
\binom{\mathbf{a}^{T}}{\overline{\mathbf{a}}^{T}} \mathbf{x}=\binom{\mathbf{a}^{T} \mathbf{p}}{\overline{\mathbf{a}}^{T} \mathbf{q}}
\end{array}
$$



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## Affine maps in 2D

## Reference

Farin and Hansford, Chapter 6

## Affine change of coordinates

We transform the point $\mathbf{x}$ into point $\mathbf{x}^{\prime}$. Note that the matrix multiplication is performed on vectors, not on points

$$
\begin{gathered}
\mathbf{v}=\mathbf{x}-\mathbf{o} \\
\mathbf{v}^{\prime}=A \mathbf{v} \\
\mathbf{x}^{\prime}=\mathbf{p}+\mathbf{v}^{\prime}
\end{gathered}
$$

In total

$$
\mathbf{x}^{\prime}=\mathbf{p}+A(\mathbf{x}-\mathbf{o})
$$

We may go back by

$$
\mathbf{x}=\mathbf{o}+A^{-1}\left(\mathbf{x}^{\prime}-\mathbf{p}\right)
$$

## Affine maps in 2D

## Translations and rotations

- Translation: $\mathbf{x}^{\prime}=\mathbf{p}+(\mathbf{x}-\mathbf{o})$
- Rotation: $\mathbf{x}^{\prime}-\mathbf{r}=R_{\alpha}(\mathbf{x}-\mathbf{r})$



## Affine maps in 2D

## Mirrors and compositions

- Mirror:

$$
\begin{gathered}
\mathbf{p}=\frac{1}{2}\left(\mathbf{x}+\mathbf{x}^{\prime}\right) \\
\mathbf{x}^{\prime}=2 \mathbf{p}-\mathbf{x}
\end{gathered}
$$

- Compositions:

$$
\begin{gathered}
\mathbf{x}^{\prime}=\mathbf{o}^{\prime}+A(\mathbf{x}-\mathbf{o}) \\
\mathbf{x}^{\prime \prime}=\mathbf{o}^{\prime \prime}+A^{\prime}\left(\mathbf{x}^{\prime}-\mathbf{o}^{\prime}\right) \\
\hline \mathbf{x}^{\prime \prime}=\mathbf{o}^{\prime \prime}+A^{\prime} A(\mathbf{x}-\mathbf{o})
\end{gathered}
$$



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## Conic sections

## Reference <br> Juan de Buegos (2000), Capítulo 11

## Conic sections

The circle, the ellipse, the parabola, and the hyperbola are all curves stemming from a section of a cone.


## Conic sections

## Conic sections

They are all second order curves

$$
\underbrace{a x_{1}^{2}+b x_{2}^{2}+c x_{1} x_{2}}_{2 \text { nd order }}+\underbrace{d x_{1}+e x_{2}}_{1 \text { st order }}+\underbrace{f}_{\text {oth order }}=0
$$

By renaming the coefficients, we may rewrite it as

$$
\begin{gathered}
a_{11} x_{1}^{2}+a_{22} x_{2}^{2}+2 a_{12} x_{1} x_{2}+2 b_{1} x_{1}+2 b_{2} x_{2}+c=0 \\
\left(\begin{array}{ll}
x_{1} & x_{2}
\end{array}\right)\left(\begin{array}{ll}
a_{11} & a_{12} \\
a_{12} & a_{22}
\end{array}\right)\binom{x_{1}}{x_{2}}+2\left(\begin{array}{ll}
b_{1} & b_{2}
\end{array}\right)\binom{x_{1}}{x_{2}}+c=0 \\
\mathbf{x}^{T} A \mathbf{x}+2 B \mathbf{x}+c=0
\end{gathered}
$$

Compare this to the more widely known equation of the parabola $y=a x^{2}+b x+c$. Finally, we can write it in a very compact form

$$
\tilde{\mathbf{x}}^{T} M \tilde{\mathbf{x}}=\left(\begin{array}{lll}
x_{1} & x_{2} & 1
\end{array}\right)\left(\begin{array}{ccc}
a_{11} & a_{12} & b_{1} \\
a_{12} & a_{22} & b_{2} \\
b_{1} & b_{2} & c
\end{array}\right)\left(\begin{array}{c}
x_{1} \\
x_{2} \\
1
\end{array}\right)=0
$$

## Conic sections

## Definition 5.1 (Conic sections)

A conic section or conics is the locus (lugar geométrico) of all points satisfying

$$
\tilde{\mathbf{x}}^{T} M \tilde{\mathbf{x}}=0
$$

Definition 5.2 (Conic equality)
Two conics $\tilde{\mathbf{x}}^{\top} M_{1} \tilde{\mathbf{x}}=0$ and $\tilde{\mathbf{x}}^{\top} M_{2} \tilde{\mathbf{x}}=0$ are the same if

$$
M_{1}=k M_{2}
$$

for some real number $k$.

## Definition 5.3 (Degenerate and ordinary conics)

A conic section is degenerate if

$$
\operatorname{det}\{M\}=0
$$

A conic section is ordinary, if it is not degenerate.

## Conic sections

## Examples of ordinary conics

Circumphere $\quad \frac{x^{2}}{r^{2}}+\frac{y^{2}}{r^{2}}=1$
Ellipse

$$
\frac{x^{2}}{a^{2}}+\frac{y^{2}}{b_{2}^{2}}=1
$$

Hyperbola

$$
\frac{x^{2}}{a^{2}}-\frac{y^{2}}{b^{2}}=1
$$

Parabola

$$
y^{2}=2 p x
$$

## Examples of degenerate conics

Two lines
Two lines
Two lines (superposed)
Two complex lines

$$
\begin{aligned}
& x^{2}-y^{2}=(x-y)(x+y)=0 \\
& x^{2}-4=(x-2)(x+2)=0 \\
& x^{2}=0 \\
& x^{2}+y^{2}=(x-i y)(x+i y)=0
\end{aligned}
$$

## Intersection of a conics and a line

## Intersection of a conics and a line

Consider the parametric equation of a line in homogeneous coordinates

$$
\tilde{\mathbf{I}}(t)=\left(\begin{array}{c}
l_{1}(t) \\
l_{2}(t) \\
1
\end{array}\right)=\left(\begin{array}{c}
p_{1}+t v_{1} \\
p_{2}+t v_{2} \\
1
\end{array}\right)=\left(\begin{array}{c}
p_{1} \\
p_{2} \\
1
\end{array}\right)+t\left(\begin{array}{c}
v_{1} \\
v_{2} \\
0
\end{array}\right)=\tilde{\mathbf{p}}+t \tilde{\mathbf{v}}
$$

We need to find a point in the line (i.e., $t$ ) such that

$$
\begin{gathered}
\tilde{\mathbf{l}}(t)^{T} M \tilde{\mathbf{l}}(t)=0 \\
(\tilde{\mathbf{p}}+t \tilde{\mathbf{v}})^{T} M(\tilde{\mathbf{p}}+t \tilde{\mathbf{v}})=0 \\
\tilde{\mathbf{v}}^{T} M \tilde{\mathbf{v}} t^{2}+2 \tilde{\mathbf{v}}^{T} M \tilde{\mathbf{p}} t+\tilde{\mathbf{p}}^{T} M \tilde{\mathbf{p}}=0
\end{gathered}
$$

This is a second order equation in $t$. If there is no solution, then the line does not intersect the conics. If there is only 1 solution, then the line is tangent to the conics. If there are 2 solutions, then the line intersects the conics (the line is secant to the conics, secante).

## Reduced equation of a conics

## Reduced equation of a conics

Let $\lambda_{1}$ and $\lambda_{2}$ be the eigenvalues of $A=\left(\begin{array}{ll}a_{11} & a_{12} \\ a_{12} & a_{22}\end{array}\right)$. Then, there exists a basis in which the conics can be expressed as

| $\lambda_{1} \neq 0, \lambda_{2} \neq 0$ | $\lambda_{1} x^{2}+\lambda_{2} y^{2}+\frac{\operatorname{det}\{M\}}{\operatorname{det}\{A\}}=0$ | Ellipses, hyperbolas, <br> pairs of intersecting lines. |
| :---: | :---: | :--- |
| $\lambda_{1}=0, \lambda_{2} \neq 0$ <br> $\operatorname{det}\{M\} \neq 0$ | $y^{2}=2 \sqrt{-\frac{\operatorname{det}\{M\}}{\lambda_{2}^{3}}} x$ | Parabolas |
| $\lambda_{1}=0, \lambda_{2} \neq 0$ <br> $\operatorname{det}\{M\}=0$ | Pairs of parallel lines |  |

## General classification of conics

## Definition 5.4 (Signature of a quadratic form)

Consider a quadratic form $Q(\mathbf{x})=\mathbf{x}^{\top} A \mathbf{x}$ and its diagonalization such that

$$
Q(\mathbf{y})=\lambda_{1} y_{1}^{2}+\lambda_{2} y_{2}^{2}+\ldots+\lambda_{n} y_{n}^{2}
$$

The signature of $Q(\mathbf{x})$ is $\left(n_{0}, n_{+}, n_{-}\right)$where $n_{0}$ is the number of null $\lambda$ coefficients, $n_{+}$the number of positve $\lambda$ coefficients, and $n_{-}$the number of negative $\lambda$ coefficients.

## Theorem 5.1

The signature of a quadratic form is invariant to changes of basis, i.e., it only depends on $Q$.

## Definition 5.5 (Signature of a matrix)

The signature of a symmetric matrix is the signature of its associated quadratic form.

## General classification of conics

General classification of conics

| A | M | Conics |
| :---: | :---: | :---: |
| $\operatorname{det}\{A\}>0$ | $\begin{gathered} \operatorname{Sig}\{M\}=(0,1,2) \text { or }(0,2,1) \\ \operatorname{Sig}\{M\}=(0,3) \text { or }(0,3,0) \\ \operatorname{Det}\{M\}=0 \end{gathered}$ | (Real) Ellipse <br> Empty set (or imaginary ellipse) <br> A point (or the intersection of two imaginary lines) |
| $\operatorname{det}\{A\}<0$ | $\begin{aligned} & \operatorname{det}\{M\} \neq 0 \\ & \operatorname{det}\{M\}=0 \end{aligned}$ | Hyperbola <br> Two secant (real) lines |
| $\operatorname{det}\{A\}=0$ | $\begin{aligned} & \operatorname{det}\{M\} \neq 0 \\ & \operatorname{det}\{M\}=0 \end{aligned}$ | Parabola <br> Two parallel (real) lines |

## Geometric transformations

## Geometric transformations

- Shift: Shift the center to $\hat{\mathbf{c}}=\left(c_{1}, c_{2}, 0\right)$

$$
(\tilde{\mathbf{x}}-\hat{\mathbf{c}})^{T} M_{1}(\tilde{\mathbf{x}}-\hat{\mathbf{c}})=0
$$

- Rotate: Rotate the conics with a rotation matrix $R$ :

$$
\begin{aligned}
& (R \tilde{\mathbf{x}})^{T} M_{1}(R \tilde{\mathbf{x}})=0 \\
& \tilde{\mathbf{x}}^{T}\left(R^{T} M_{1} R\right) \tilde{\mathbf{x}}=0
\end{aligned}
$$

with $R=\left(\begin{array}{ccc}\cos (\alpha) & -\sin (\alpha) & 0 \\ \sin (\alpha) & \cos (\alpha) & 0 \\ 0 & 0 & 1\end{array}\right)$.

## Ellipse

## Ellipse

Reduced equation: $\quad \frac{x^{2}}{a^{2}}+\frac{y^{2}}{b^{2}}=1$
Parametric equation:

$$
\begin{aligned}
& x=a \cos t \\
& y=b \sin t \\
& t \in[0,2 \pi) \\
& d\left(F, F^{\prime}\right)=2 c \\
& \text { where } \\
& a^{2}+b^{2}=c^{2}
\end{aligned}
$$

$$
\text { Interfocal distance: } \quad d\left(F, F^{\prime}\right)=2 c
$$



## Hyperbola

## Hyperbola

Reduced equation:
Parametric equation:

$$
\begin{aligned}
& \frac{x^{2}}{a^{2}}-\frac{y^{2}}{b^{2}}=1 \\
& x= \pm a \cosh t \\
& y=b \sinh t \\
& t \in \mathbb{R} \\
& d\left(F, F^{\prime}\right)=2 c \\
& \text { where } \\
& a^{2}+b^{2}=c^{2}
\end{aligned}
$$

Interfocal distance:


## (Calculus note)

$$
\begin{array}{c|c}
\cos x=\frac{e^{i x}+e^{-i x}}{2} & \cosh x=\frac{e^{x}+e^{-x}}{2} \\
\sin x=\frac{e^{i x}-e^{-i x}}{2} & \sinh x=\frac{e^{x}-e^{-x}}{2} \\
\cos ^{2} x+\sin ^{2} x=1 & \cosh ^{2} x-\sinh ^{2} x=1
\end{array}
$$

## Parabola

## Parabola

Reduced equation:

$$
\begin{aligned}
& y^{2}=2 p x \\
& x=\frac{t^{2}}{2 p} \\
& y=t \\
& t \in \mathbb{R}
\end{aligned}
$$



## Outline

(9) Linear algebra applications in geometry

- Local and global coordinates
- Points and vectors
- Lines in 2D
- Affine maps in 2D
- Conic sections in 2D
- 3D Geometry
- Quadrics in 3D


## Cross product

## Reference

## Farin and Hansford, Chapter 10

## Cross product

The cross product is defined for 3D vectors as

$$
\mathbf{u}=\mathbf{v} \times \mathbf{w}=\left|\begin{array}{lll}
\mathbf{e}_{1} & \mathbf{e}_{2} & \mathbf{e}_{3} \\
v_{1} & v_{2} & v_{3} \\
w_{1} & w_{2} & w_{3}
\end{array}\right|
$$

Properties:

$$
\begin{aligned}
& \mathbf{u} \perp \mathbf{v} \text { and } \mathbf{u} \perp \mathbf{w} \\
& \|\mathbf{v} \times \mathbf{w}\|^{2}=\|\mathbf{v}\|\|\mathbf{w}\|-(\mathbf{v} \cdot \mathbf{w})^{2} \\
& \mathbf{v} \times(c \mathbf{v})=\mathbf{0} \\
& \mathbf{v} \times(c \mathbf{w})=(c \mathbf{v}) \times \mathbf{w}=c(\mathbf{v} \times \mathbf{w}) \\
& \mathbf{w} \times \mathbf{v}=-\mathbf{v} \times \mathbf{w} \\
& \mathbf{u} \times(\mathbf{v}+\mathbf{w})=\mathbf{u} \times \mathbf{v}+\mathbf{u} \times \mathbf{w} \\
& \mathbf{u} \times(\mathbf{v} \times \mathbf{w}) \neq(\mathbf{u} \times \mathbf{v}) \times \mathbf{w}
\end{aligned}
$$



## Cross product

## Example

$$
\begin{aligned}
\mathbf{u} & =\mathbf{e}_{1} \times \mathbf{e}_{2}=\left|\begin{array}{ccc}
\mathbf{e}_{1} & \mathbf{e}_{2} & \mathbf{e}_{3} \\
1 & 0 & 0 \\
0 & 1 & 0
\end{array}\right|=\mathbf{e}_{3} \\
\mathbf{u} & =\mathbf{e}_{2} \times \mathbf{e}_{1}=\left|\begin{array}{ccc}
\mathbf{e}_{1} & \mathbf{e}_{2} & \mathbf{e}_{3} \\
0 & 1 & 0 \\
1 & 0 & 0
\end{array}\right|=-\mathbf{e}_{3}
\end{aligned}
$$

## Cross product

## Coordinate systems

- Right-handed:

$$
\begin{aligned}
& \mathbf{x} \times \mathbf{y}=\mathbf{z} \\
& \mathbf{y} \times \mathbf{z}=\mathbf{x} \\
& \mathbf{z} \times \mathbf{x}=\mathbf{y}
\end{aligned}
$$

- Left-handed:
$\mathbf{x} \times \mathbf{y}=-\mathbf{z}$
$y \times z=x$
$\mathbf{z} \times \mathbf{x}=-\mathbf{y}$



## Cross product

## Area of parallelogram

The norm of $\mathbf{v} \times \mathbf{w}$ is the area of the parallelogram formed by $\mathbf{u}$ and $\mathbf{v}$ and is equal to:

$$
A=\|\mathbf{v} \times \mathbf{w}\|=\|\mathbf{v}\|\|\mathbf{w}\| \sin (\theta)
$$



## Lines

## Parametric equation of a line

A line is defined in 3D (and nD ) by two points or a point and a vector

- Given two points:

$$
\overline{\mathbf{I}(t)=\mathbf{p}+t(\mathbf{q}-\mathbf{p}) \quad t \in \mathbb{R}, ~}
$$

- Given point and vector:

$$
\overline{\mathbf{l}(t)=\mathbf{p}+t \mathbf{v} \quad t \in \mathbb{R}}
$$

Giving a point and a perpendicular vector does no longer work.

## Planes

## Implicit equation of a planes

A plane is defined in 3D by a point and a perpendicular vector

- Given a point and the normal direction: $\overline{\mathbf{n} \cdot(\mathbf{x}-\mathbf{p})=0}$

In 3D:

$$
\begin{gathered}
\left(n_{1}, n_{2}, n_{3}\right) \cdot\left(x_{1}-p_{1}, x_{2}-p_{2}, x_{3}-p_{3}\right)=0 \Rightarrow \\
A x_{1}+B x_{2}+C x_{3}+D=0
\end{gathered}
$$

The absolute value of $D$ in the implicit equation is the distance of the plane to the coordinate system origin.


## Hyperplanes

## Hyperplanes

A hyperplane of $\mathbb{R}^{n}$ is an affine space of a dimension $n-1$. For instance

| $\mathbb{R}^{n}$ | Dimension | Dimension of hyperplane | Hyperplane name |
| :--- | :---: | :---: | :--- |
| $\mathbb{R}^{2}$ | $2 D$ | 1 | Line |
| $\mathbb{R}^{3}$ | $3 D$ | 2 | Plane |
| $\mathbb{R}^{n}$ | $n D$ | $n-1$ | Hyperplane |

All hyperplanes are defined by a point (p) and a normal vector (n)

$$
\mathbf{n} \cdot(\mathbf{x}-\mathbf{p})=0
$$

## Distance of a point to a plane (hyperplane)

The distance between a point $\mathbf{r}$ and a plane (or hyperplane) is given by

$$
d=\frac{\mathbf{n} \cdot(\mathbf{r}-\mathbf{p})}{\|\mathbf{n}\|}
$$

## Planes

## Parametric equation of a plane

A plane can also be defined in 3D (and nD) by a point and two in-plane vectors

- Given a point and two in-plane vectors: $\overline{P(s, t)=\mathbf{p}+s \mathbf{v}+t \mathbf{w} \quad \forall s, t \in \mathbb{R}, ~(1)}$
- Given three points:

$$
\overline{P(s, t)=\mathbf{p}+s(\mathbf{q}-\mathbf{p})+t(\mathbf{r}-\mathbf{p}) \quad \forall s, t \in \mathbb{R}, ~}
$$



## Scalar triple product

## Scalar triple product

The volume of a parallelepiped can be measured with the scalar triple product

$$
V=\mathbf{u} \cdot(\mathbf{v} \times \mathbf{w})
$$

Properties:

$$
\mathbf{u} \cdot(\mathbf{v} \times \mathbf{w})=\mathbf{v} \cdot(\mathbf{u} \times \mathbf{w})=\mathbf{w} \cdot(\mathbf{v} \times \mathbf{u})
$$



## Distance between two lines

Distance between two lines
Given two lines in parametric form

$$
\mathbf{I}_{1}\left(s_{c}\right)=\mathbf{p}_{0}+s_{c} \mathbf{u} \mathbf{I}_{2}\left(t_{c}\right)=\mathbf{q}_{0}+t_{c} \mathbf{v}
$$

The distance between the two lines is the length of the vector $\mathbf{w}_{c}$ that is perpendicular to both lines. $\mathbf{w}_{c}$ is defined by two points: one in line $1\left(\mathbf{x}_{1}\right)$ and another one in line $2\left(\mathbf{x}_{2}\right)$ :

$$
\mathbf{w}_{c}=\mathbf{x}_{2}-\mathbf{x}_{1}=\mathbf{q}_{0}+t_{c} \mathbf{v}-\left(\mathbf{p}_{0}+s_{c} \mathbf{u}\right)
$$

The conditions on $\mathbf{w}_{c}$ are:

$$
\mathbf{w}_{c} \cdot \mathbf{u}=0 \text { and } \mathbf{w}_{c} \cdot \mathbf{v}=0
$$



After reorganizing the terms

$$
\left(\begin{array}{cc}
\|\mathbf{u}\|^{2} & -\mathbf{u} \cdot \mathbf{v} \\
\mathbf{u} \cdot \mathbf{v} & \|\mathbf{v}\|^{2}
\end{array}\right)\binom{s_{c}}{t_{c}}=\binom{\left(\mathbf{p}_{0}-\mathbf{q}_{0}\right) \cdot \mathbf{u}}{\left(\mathbf{p}_{0}-\mathbf{q}_{0}\right) \cdot \mathbf{v}}
$$

## Intersection of two lines

## Intersection of two lines

The two lines in the previous slide intersect if $\mathbf{x}_{1}=\mathbf{x}_{2}$. We also note that the two lines intersect if $\mathbf{u}, \mathbf{v}$ and $\mathbf{p}_{0}-\mathbf{q}_{0}$ are in the same plane, or what is the same they are linearly dependent

$$
\left|\left(\begin{array}{lll}
\mathbf{u} & \mathbf{v} & \mathbf{p}_{0}-\mathbf{q}_{0}
\end{array}\right)\right|=0
$$

## Intersection of a line and a plane

## Intersection of a line and a plane

- Parametric line, implicit plane:

$$
\begin{gathered}
\mathbf{l}(t)=\mathbf{p}+t \mathbf{v} \\
\mathbf{n} \cdot(\mathbf{x}-\mathbf{q})=0
\end{gathered}
$$

For the intersection we need to find $t$ such that

$$
\mathbf{n} \cdot(\mathbf{p}+t \mathbf{v}-\mathbf{q})=0
$$

whose solution is

$$
\begin{gathered}
t=\frac{\mathbf{n} \cdot(\mathbf{q}-\mathbf{p})}{\mathbf{n} \cdot \mathbf{v}} \\
\mathbf{x}=\mathbf{p}+\frac{\mathbf{n} \cdot(\mathbf{q}-\mathbf{p})}{\mathbf{n} \cdot \mathbf{v}} \mathbf{v}
\end{gathered}
$$



## Intersection of a line and a plane

## Intersection of a line and a plane

- Parametric line, parametric plane:

$$
\begin{gathered}
\mathbf{l}(t)=\mathbf{p}+t \mathbf{v} \\
\mathbf{P}\left(t_{1}, t_{2}\right)=\mathbf{q}+t_{1} \mathbf{u}+t_{2} \mathbf{w}
\end{gathered}
$$

We need to find $t, t_{1}$ and $t_{2}$ such that

$$
\mathbf{p}+t \mathbf{v}=\mathbf{q}+t_{1} \mathbf{u}+t_{2} \mathbf{w}
$$

Reorganizing the terms:

$$
\left(\begin{array}{lll}
\mathbf{u} & \mathbf{w} & -\mathbf{v}
\end{array}\right)\left(\begin{array}{c}
t_{1} \\
t_{2} \\
t
\end{array}\right)=\mathbf{p}-\mathbf{q}
$$



## Intersection of a line and a triangle

## Intersection of a line and a triangle

- Parametric line, 3 points of a triangle:

$$
\begin{aligned}
& \mathbf{l}(t)=\mathbf{p}+t \mathbf{v} \\
& \mathbf{P}\left(t_{1}, t_{2}\right)= \mathbf{p}_{1}+t_{1}\left(\mathbf{p}_{2}-\mathbf{p}_{1}\right)+t_{2}\left(\mathbf{p}_{3}-\mathbf{p}_{1}\right) \\
& t_{1}, t_{2} \in[0,1], t_{1}+t_{2} \leq 1
\end{aligned}
$$

We need to find $t, t_{1}$ and $t_{2}$ such that

$$
\mathbf{p}+t \mathbf{v}=\mathbf{p}_{1}+t_{1}\left(\mathbf{p}_{2}-\mathbf{p}_{1}\right)+t_{2}\left(\mathbf{p}_{3}-\mathbf{p}_{1}\right)
$$

Reorganizing the terms:

$$
\left(\begin{array}{lll}
\mathbf{p}_{2}-\mathbf{p}_{1} & \mathbf{p}_{3}-\mathbf{p}_{1} & -\mathbf{v}
\end{array}\right)\left(\begin{array}{c}
t_{1} \\
t_{2} \\
t
\end{array}\right)=\mathbf{p}-\mathbf{p}_{1}
$$



The intersection point is within the triangle if $t_{1}, t_{2} \in[0,1], t_{1}+t_{2} \leq 1$.

## Reflection

## Reflection

- Reflection:

This situation is encountered, for instance, in reflected light rays. By inspecting the figure we note that

$$
\mathbf{n} \cdot \mathbf{v}=-\mathbf{n} \cdot \mathbf{v}^{\prime}
$$

On the other side, it must also be

$$
c \mathbf{n}=\mathbf{v}^{\prime}-\mathbf{v}
$$

We have two unknowns $c$ and $\mathbf{v}$ and two equations. After some manipulation we reach


$$
\mathbf{v}^{\prime}=\mathbf{v}-2\left(\mathbf{n} \cdot \mathbf{n}^{T}\right) \mathbf{v}
$$

## Intersection of three planes

## Intersection of three planes

- Implicit equations:

For each of the planes, we have

$$
\begin{aligned}
& \mathbf{n}_{1} \cdot\left(\mathbf{x}-\mathbf{p}_{1}\right)=0 \Rightarrow \mathbf{n}_{1}^{T} \mathbf{x}=\mathbf{n}_{1}^{T} \mathbf{p}_{1} \\
& \mathbf{n}_{2} \cdot\left(\mathbf{x}-\mathbf{p}_{2}\right)=0 \Rightarrow \mathbf{n}_{2}^{T} \mathbf{x}=\mathbf{n}_{2}^{T} \mathbf{p}_{2} \\
& \mathbf{n}_{3} \cdot\left(\mathbf{x}-\mathbf{p}_{3}\right)=0 \Rightarrow \mathbf{n}_{3}^{T} \mathbf{x}=\mathbf{n}_{3}^{T} \mathbf{p}_{3}
\end{aligned}
$$

Gathering all together

$$
\left(\begin{array}{l}
\mathbf{n}_{1}^{T} \\
\mathbf{n}_{2}^{T} \\
\mathbf{n}_{3}^{T}
\end{array}\right) \mathbf{x}=\left(\begin{array}{c}
\mathbf{n}_{1}^{T} \mathbf{p}_{1} \\
\mathbf{n}_{2}^{T} \mathbf{p}_{2} \\
\mathbf{n}_{3}^{T} \mathbf{p}_{3}
\end{array}\right)
$$

In non-degenerate situations, this equation system has a unique solution that is the intersection point. Otherwise, the planes
 may intersect in one line, two lines, three lines, or even in a plane (if the three planes are the same plane).

## Intersection of two planes

## Intersection of two planes

- Implicit equations:

For each of the planes, we have

$$
\begin{gathered}
\mathbf{n} \cdot\left(\mathbf{x}-\mathbf{p}_{1}\right)=0 \Rightarrow \mathbf{n}^{T} \mathbf{x}=\mathbf{n}^{T} \mathbf{p}_{1} \\
\mathbf{m} \cdot\left(\mathbf{x}-\mathbf{p}_{2}\right)=0 \Rightarrow \mathbf{m}^{T} \mathbf{x}=\mathbf{m}^{T} \mathbf{p}_{2}
\end{gathered}
$$

The two planes intersect in a line of the form

$$
\mathbf{I}(t)=\mathbf{p}+t(\mathbf{n} \times \mathbf{m})
$$

To find $\mathbf{p}$ we solve the equation system

$$
\left(\begin{array}{c}
\mathbf{n}^{T} \\
\mathbf{m}^{T} \\
(\mathbf{n} \times \mathbf{m})^{T}
\end{array}\right) \mathbf{x}=\left(\begin{array}{c}
\mathbf{n}^{T} \mathbf{p}_{1} \\
\mathbf{m}^{T} \mathbf{p}_{2} \\
0
\end{array}\right)
$$



## Outline

(9) Linear algebra applications in geometry

- Local and global coordinates
- Points and vectors
- Lines in 2D
- Affine maps in 2D
- Conic sections in 2D
- 3D Geometry
- Quadrics in 3D


## Quadrics

## Reference

Juan de Buegos (2000), Capítulo 12

## Quadrics

Quadrics are 3D surfaces that meet a second order equation.

$z-x^{2}+y^{2}$


Hyperbolic Paraboloid
$z=x^{2}-y^{2}$


Ellipsoid
$x^{2}+y^{2}+z^{2}=1$


Hyperboloid - One Sheet
$x^{2}+y^{2}-z^{2}-1$


Hyperboloid - Two Sheets $x^{2}-y^{2}-z^{2}=1$

Quadrics in the Wikipedia

## Ellipsoid

## Ellipsoid

Reduced equation:

$$
\begin{gathered}
\frac{x^{2}}{a^{2}}+\frac{y^{2}}{b^{2}}+\frac{z^{2}}{c^{2}}=1 \\
x=a \cos u \sin v \\
y=b \sin u \sin v \\
z=c \cos v \\
u, v \in[0,2 \pi)
\end{gathered}
$$

Parametric equation: $\quad y=b \sin u \sin v$

Cuts along $X, Y$ and $Z$ are ellipses.


## Hyperboloid of one sheet

## Hyperboloid of one sheet

Reduced equation: $\quad \frac{x^{2}}{a^{2}}+\frac{y^{2}}{b^{2}}-\frac{z^{2}}{c^{2}}=1$

$$
x=a \sqrt{1+u^{2}} \cos v
$$

Parametric equation: $\quad y=b \sqrt{1+u^{2}} \sin v$
$z=c u$
$x=a \cosh u \cos v$
Parametric equation: $\quad y=b \cosh u \sin v$

$$
z=c \sinh u
$$

$$
v \in[0,2 \pi), u \in \mathbb{R}
$$

Cuts along $X$ and $Y$ are hyperbolas.


Cuts along $Z$ are ellipses.

## Hyperboloid of two sheets

Hyperboloid of two sheets

Reduced equation:

$$
\begin{gathered}
\frac{x^{2}}{a^{2}}+\frac{y^{2}}{b^{2}}-\frac{z^{2}}{c^{2}}=-1 \\
x=a \sinh u \cos v
\end{gathered}
$$

Parametric equation:

$$
\begin{aligned}
& y=b \sinh u \sin v \\
& z=c \cosh u \\
& v \in[0,2 \pi), u \in \mathbb{R}
\end{aligned}
$$

Cuts along $X$ and $Y$ are hyperbolas. Cuts along $Z$ are ellipses.


## Elliptic paraboloid

Elliptic paraboloid
Reduced equation: $\quad \frac{x^{2}}{a^{2}}+\frac{y^{2}}{b^{2}}-\frac{z}{c}=0$

$$
x=a \sqrt{u} \cos v
$$

Parametric equation: $\quad y=b \sqrt{u} \sin v$

$$
z=c u
$$

$$
v \in[0,2 \pi), u \in[0, \infty)
$$

Cuts along $X$ and $Y$ are parabolas.
 Cuts along $Z$ are ellipses.

## Hyperbolic paraboloid

Hyperbolic paraboloid
Reduced equation:

$$
\begin{aligned}
& \frac{x^{2}}{a^{2}}-\frac{y^{2}}{b^{2}}-\frac{z}{c}=0 \\
& x=a \sqrt{u} \cosh v \\
& y=b \sqrt{u} \sinh v \\
& z=c u \\
& u, v \in \mathbb{R}
\end{aligned}
$$

Parametric equation: $\quad y=b \sqrt{u} \sinh v$

Cuts along $Y$ are parabolas.
Cuts along $Z$ are hyperbolas.


## Cone

## Cone

Reduced equation:

$$
\begin{aligned}
& \frac{x^{2}}{a^{2}}+\frac{y^{2}}{b^{2}}-\frac{z^{2}}{c^{2}}=0 \\
& x=a u \cos v \\
& y=b u \sin v \\
& z=c u \\
& v \in[0,2 \pi), u \in \mathbb{R}
\end{aligned}
$$

Parametric equation: $\quad y=b u \sin v$

Cuts along $Y$ are parabolas.
Cuts along $Z$ are ellipses.


## Elliptic cylinder

## Elliptic cylinder

Reduced equation:

$$
\begin{aligned}
& \frac{x^{2}}{a^{2}}+\frac{y^{2}}{b^{2}}=1 \\
& x=a \cos v \\
& y=b \sin v \\
& z=u \\
& v \in[0,2 \pi), u \in \mathbb{R}
\end{aligned}
$$

Parametric equation: $\quad y=b \sin v$

Cuts along $X$ and $Y$ are pairs of lines. Cuts along $Z$ are ellipses.

## Hyperbolic cylinder

## Hyperbolic cylinder

Reduced equation:

$$
\begin{gathered}
\frac{x^{2}}{a^{2}}-\frac{y^{2}}{b^{2}}=1 \\
x=a \cosh v \\
y=b \sinh v \\
z=u \\
u, v \in \mathbb{R}
\end{gathered}
$$

Parametric equation: $\quad y=b \sinh v$

Cuts along $X$ and $Y$ are pairs of lines. Cuts along $Z$ are hyperbolas.


## Parabolic cylinder

## Parabolic cylinder

$$
\begin{array}{lc}
\text { Reduced equation: } & \frac{x^{2}}{a^{2}}-\frac{y}{b}=0 \\
& x=a u \\
\text { Parametric equation: } & y=b u^{2} \\
& z=v \\
& u, v \in \mathbb{R}
\end{array}
$$

Cuts along $X$ and $Y$ are pairs of lines or single lines.
Cuts along $Z$ are parabolas.


## Quadrics

## Definition 7.1

Quadrics All quadrics can be written as

$$
\begin{gathered}
\sum_{i, j=1}^{3} a_{i j} x_{i} x_{j}+2 \sum_{i=1}^{3} b_{i} x_{i}+c=0 \\
\tilde{\mathbf{x}}^{T} M \tilde{\mathbf{x}}=0
\end{gathered}
$$

with $a_{i j}=a_{j i}$ and

$$
M=\left(\begin{array}{cccc}
a_{11} & a_{12} & a_{13} & b_{1} \\
a_{12} & a_{22} & a_{23} & b_{2} \\
a_{13} & a_{23} & a_{33} & b_{3} \\
b_{1} & b_{2} & b_{3} & c
\end{array}\right) \text { and } \tilde{\mathbf{x}}=\left(\begin{array}{c}
x_{1} \\
x_{2} \\
x_{3} \\
1
\end{array}\right)
$$

## Quadrics

## Definition 7.2 (Quadrics equality)

Two quadrics $\tilde{\mathbf{x}}^{T} M_{1} \tilde{\mathbf{x}}=0$ and $\tilde{\mathbf{x}}^{T} M_{2} \tilde{\mathbf{x}}=0$ are the same if

$$
M_{1}=k M_{2}
$$

for some real number $k$.

## Definition 7.3 (Degenerate or ordinary quadrics)

A quadrics is degenerate if $\operatorname{det}\{M\}=0$ (e.g., cones, cylinders and pairs of planes). It is ordinary if it is not degenerate (e.g., ellipsoids, paraboloids, hyperboloids)

## Examples of degenerate quadrics

$$
\begin{array}{ll}
x^{2}-y^{2}=0=(x-y)(x+y) & \text { A pair of planes } \\
x^{2}+y^{2}=0=(x-i y)(x+i y) & \text { A pair of imaginary planes } \\
x^{2}-1=0=(x-1)(x+1) & \text { A pair of planes } \\
x^{2}+y^{2}-25=0 & \text { Cylinder of radius } 5
\end{array}
$$

## General classification of quadrics

## General classification of quadrics

Let $\lambda_{1}, \lambda_{2}$ and $\lambda_{3}$ be the eigenvalues of $A$. Then, there exists a basis such that the reduced equation of the quadrics is

| Condition | Quadrics |
| :---: | :--- |
| $\lambda_{1} \neq 0, \lambda_{2} \neq 0, \lambda_{3} \neq 0$ | $\lambda_{1} x^{2}+\lambda_{2} y^{2}+\lambda_{3} z^{2}+\frac{\operatorname{det}\{M\}}{\operatorname{det}\{A\}}=0$ <br> Ellipsoids, hyperboloids and cones |
| $\lambda_{1} \neq 0, \lambda_{2} \neq 0, \lambda_{3}=0$ <br> $\operatorname{det}\{M\} \neq 0$ | $\lambda_{1} x^{2}+\lambda_{2} y^{2}=2 \sqrt{-\frac{\operatorname{det}\{M\}}{\lambda_{1} \lambda_{2}}} z$ <br> Paraboloid |
| $\lambda_{1} \neq 0, \lambda_{2} \neq 0, \lambda_{3}=0$ <br> $\operatorname{det}\{M\}=0$ | $\lambda_{1} x^{2}+\lambda_{2} y^{2}=k$ <br> Elliptical cylinder |
| $\lambda_{1}=0, \lambda_{2} \neq 0, \lambda_{3}=0$ <br> $\operatorname{Rank}\{M\}=3$ | $y^{2}=2 q x$ Parabolic cylinder |
| $\lambda_{1}=0, \lambda_{2} \neq 0, \lambda_{3}=0$ <br> $\operatorname{Rank}\{M\}<3$ | $y^{2}=k$ Pair of planes |

## Outline

(9) Linear algebra applications in geometry

- Local and global coordinates
- Points and vectors
- Lines in 2D
- Affine maps in 2D
- Conic sections in 2D
- 3D Geometry
- Quadrics in 3D


# Chapter 10. Abstract algebra 

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## Outline

(10) Abstract algebra

- Sets
- Relations and functions
- Partitions and equivalence relationships
- Binary operations
- Groups and subgroups
- Homomorphisms and isomorphisms
- Algebraic structures


## References


J.B. Fraleigh. A first course in Abstract Algebra. Pearson, 7th Ed. (2002)

## Outline

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- Sets
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## Sets

## Definition 1.1 (Set)

A set is a well-defined collection of elements. We denote the different elements as $a \in S$.

## Definition 1.2 (Empty set)

The only set without any element is the empty set ( $\emptyset$ ).

## Describing sets

We may provide the elements of a set:

- Intensional definition: by giving a property they all meet (e.g., even numbers from 1 to 10 )
- Extensional definition: by listing all the elements in the set (e.g., $\{2,4,6,8,10\}$ ). The order in which the different elements are written has no meaning.


## Sets

## Definition 1.3 (Subset and proper subset)

$B$ is subset of $A$ (denoted $B \subseteq A$ or $A \supseteq B$ ) if all the elements of $B$ are also elements of $A$. $B$ is a proper subset of $A$ if $B$ is a subset of $A$ and $B$ is different from $A(B \subset A$ or $A \supset B)$.

## Properties

- $A$ is an improper subset of $A$.
- $\emptyset$ is a proper subset of $A$.


## Definition 1.4 (Power set (Partes de un conjunto))

The set of all subsets of a set $A$ is called the power set of $A$.

## Example

Let $A=\{1,2,3\}$ the power set of $A$ is

$$
P(A)=\{\emptyset,\{1\},\{2\},\{3\},\{1,2\},\{1,3\},\{2,3\},\{1,2,3\}\}
$$

## Sets

## Definition 1.5 (Cartesian product)

The cartesian product of the sets $A$ and $B$ is the set of all ordered pairs in which the first element comes from $A$ and the second element comes from $B$.

$$
A \times B=\{(a, b) \mid a \in A, b \in B\}
$$

Note that because of the ordered nature of the pair $A \times B \neq B \times A$.

## Example

Let $A=\{1,2,3\}$ and $B=\{4,5\}$.

$$
A \times B=\{(1,4),(1,5),(2,4),(2,5),(3,4),(3,5)\}
$$

## Definition 1.6 (Cardinality)

The cardinality of a set is the number of elements it has.

## Sets

## Definition 1.7 (Disjoint sets)

Two sets are disjoint if they do not have any element in common.

## Some useful sets

- Integer numbers: $\mathbb{Z}=\{\ldots,-2,-1,0,1,2, \ldots\},|\mathbb{Z}|=\aleph_{0}$
- Natural numbers, positive integers: $\mathbb{N}=\mathbb{Z}^{+}=\{1,2,3, \ldots\},|\mathbb{N}|=\aleph_{0}$
- Negative integers: $\mathbb{Z}^{-}=\{\ldots,-3,-2,-1\},\left|\mathbb{Z}^{-}\right|=\aleph_{0}$
- Non-null integers: $\mathbb{Z}^{*}=\mathbb{Z}-\{0\}=\{\ldots,-2,-1,1,2, \ldots\},\left|\mathbb{Z}^{*}\right|=\aleph_{0}$
- Rational numbers: $\mathbb{Q},|\mathbb{Q}|=\aleph_{0}$
- Real numbers: $\mathbb{R},|\mathbb{R}|=\aleph_{1}$
- Interval: $[0,1],|[0,1]|=\aleph_{1}$
- Complex numbers: $\mathbb{C}=\{a+b i \mid a, b \in \mathbb{R}\},|\mathbb{C}|=\aleph_{1}$


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- Algebraic structures


## Relations

## Definition 2.1 (Relation)

A relation $a R b$ is a subset of the cartesian product $A \times B$.

## Example



## Functions

## Definition 2.2 (Function)

A function $f: X \rightarrow Y$ is a relation between $X$ and $Y$ in which each $x \in X$ appears at most in one of the pairs $(x, y)$. We may write

$$
(x, y) \in f \text { or } f(x)=y
$$

The domain of $f$ is $X$, the codomain of $f$ is $Y$. The support of $f$ is the set of all those values in $X$ for which there exists a pair $(x, y)$. The range of $f$ are all values in $Y$ for which there exists at least one pair $(x, y)$.

## Example

$$
\begin{gathered}
f: \mathbb{R} \rightarrow \mathbb{R} \\
f(x)=x^{3} \\
(2,8) \in f \Leftrightarrow f(2)=8 \\
+: \mathbb{R} \times \mathbb{R} \rightarrow \mathbb{R} \\
((2,3), 5) \in+\Leftrightarrow+((2,3))=5 \Leftrightarrow 2+3=5
\end{gathered}
$$

## Classification of functions

## Definition 2.3

Functions can be classified as surjective, injective or bijective:
Surjective: A function is surjective if every point of the codomain has at least one point of the domain that maps onto it. They are also called onto functions.

Injective: A function is injective if every point of the codomain has at most one point in the domain that maps onto it. They are also called one-to-one functions.
Bijective: A function is bijective if it is injective and surjective.

surjection

injection

bijection

## Inverse function

## Definition 2.4 (Inverse function)

Consider an injective function $f: X \rightarrow Y . f^{-1}: Y \rightarrow X$ is the inverse of $f$ iff

$$
(x, y) \in f \Rightarrow(y, x) \in f^{-1}
$$

## Example

- $f(x)=x+3 \Rightarrow f^{-1}(y)=y-3$
- $f(x)=x^{3} \Rightarrow f^{-1}(y)=y^{\frac{1}{3}}$
- $f(x)=x^{2}$ is not invertible because it is not injective $(f(-2)=f(2)=4)$


## Inverse function

## Theorem 2.1

- If $f$ is invertible, its inverse is unique.
- If $f$ is bijective, so is $f^{-1}$.
- $X$ and $Y$ have the same cardinality if there exists a bijective function between the two.


## Example

Consider the following function $f: \mathbb{Z} \rightarrow \mathbb{N}$

$$
\begin{gathered}
0 \\
0
\end{gathered}-1
$$

$f$ is bijective. Consequently, $\mathbb{Z}$ has the same cardinality as $\mathbb{N}$.

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## Partition

## Definition 3.1 (Partition)

A partition of a set $S$ is a collection of non-empty subsets such that each element of $S$ belongs to one and only one subset (cell) of the partition. We denote as $\bar{x}$ the subset that contains the element $x$. All cells in a partition are disjoint to any other cell.

## Examples

- We may partition the set of natural numbers into the subset of even numbers $(\{2,4,6, \ldots\})$ and the subset of odd numbers ( $\{1,3,5, \ldots\}$ ).
- We may partition the set of integer numbers into the subset of all multiples of $3(\{\ldots,-6,-3,0,3,6, \ldots\})$, the subset of all numbers whose remainder after dividing by 3 is $1(\{\ldots,-5,-2,1,4,7, \ldots\})$, and the subset of all numbers whose remainder after dividing by 3 is $2(\{\ldots,-4,-1,2,5,8, \ldots\})$.


## Equivalence relation

## Definition 3.2 (Equivalence relation)

$R$ is an equivalence relation in $S$ if it verifies:
(1) $R$ is reflexive: $x R x$
(2) $R$ is symmetric: $x R y \Rightarrow y R x$
(3) $R$ is transitive: $x R y, y R z \Rightarrow x R z$

## Examples

(1) = is an equivalence relation.
(2) Congruence modulo n is an equivalence relation (two numbers are related if they have the same remainder after dividing by $n$ )
Example: 1 and 4 have remainder 1 after dividing by 3 . We write

$$
1 \equiv 4(\bmod 3)
$$

(3) $\forall n, m \in \mathbb{Z} \quad n R m \Leftrightarrow n m \geq 0$ is not an equivalence relationship because it is not transitive (e.g., $-3 R 0,0 R 5$ but $-3 R$ ).

## Partition and equivalence relation

## Theorem 3.1

Let $S$ be a non-empty set, and $R$ an equivalence relation defined on $S$. Then $R$ partitions $S$ with the cells

$$
\bar{a}=\{x \in S \mid x R a\}
$$

Additionally, we may define another equivalence relation ~

$$
a \sim b \Leftrightarrow \bar{a}=\bar{b}
$$

## Partition and equivalence relation

## Example

Congruence modulo 3 is an equivalence relation in $\mathbb{Z}$ (two numbers are related if they have the same remainder after dividing by 3 )

$$
\begin{aligned}
& \overline{0}=\{\ldots,-6,-3,0,3,6, \ldots\} \\
& \overline{1}=\{\ldots,-5,-2,1,4,7, \ldots\} \\
& \overline{2}=\{\ldots,-4,-1,2,5,8, \ldots\}
\end{aligned}
$$

Additionally

$$
\begin{aligned}
& \ldots=\overline{0}=\overline{3}=\overline{6}=\ldots \Rightarrow 0 \sim 3 \sim 6 \sim \ldots \\
& \ldots=\overline{1}=\overline{4}=\overline{7}=\ldots \Rightarrow 1 \sim 4 \sim 7 \sim \ldots \\
& \ldots=\overline{2}=\overline{5}=\overline{8}=\ldots \Rightarrow 2 \sim 5 \sim 8 \sim \ldots
\end{aligned}
$$

and

$$
\mathbb{Z}=\overline{0} \cup \overline{1} \cup \overline{2}
$$

## Partition and equivalence relation

## Example

Consider the cartesian product $\mathbb{Z} \times(\mathbb{Z}-\{0\})$. Let $\left(m_{1}, n_{1}\right)$ and $\left(m_{2}, n_{2}\right)$ be two ordered sets of this cartesian product. Consider now the equivalence relation

$$
\left(m_{1}, n_{1} \sim\left(m_{2}, n_{2}\right) \Leftrightarrow m_{1} n_{2}-m_{2} n_{1}=0\right.
$$

The set of rational numbers is formally defined $\mathbb{Q}$ as the set of equivalence classes of $\mathbb{Z} \times(\mathbb{Z}-\{0\})$ under the relation $\sim$.

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## Binary operations

## Introduction

## What is addition?

Let us assume that we arrive to a classroom in Mars, and that martians are learning to add. The teacher says

Gloop, poyt
and the students reply:
Bimt.
Then, the teacher says:
Ompt, gaft
and the students reply:
Poyt.
We don't know what they do but it seems that when the teacher gives two elements, students respond with another element.

## Binary operations

## Introduction (continued)

## What is addition?

This is what we do when we say "three plus four", "seven". And we may not use any two elements ("three plus apples" is not defined). We can only use elements on a given set. This is what we formally call a binary operation.

## Definition 4.1 (Binary operation)

A binary operation on a set $S$ is a function:

$$
\begin{aligned}
*: S \times S & \rightarrow S \\
*(a, b) & =a * b
\end{aligned}
$$

## Binary operations

## Examples

The following binary operations are all different:

$$
\begin{aligned}
& +: \mathbb{R} \times \mathbb{R} \rightarrow \mathbb{R} \\
& +: \mathbb{Z} \times \mathbb{Z} \rightarrow \mathbb{Z} \\
& +: \mathcal{M}_{m \times n}(\mathbb{R}) \times \mathcal{M}_{m \times n}(\mathbb{R}) \rightarrow \mathcal{M}_{m \times n}(\mathbb{R})
\end{aligned}
$$

The following is not a binary operation because it is not well defined

$$
+: \mathcal{M}(\mathbb{R}) \times \mathcal{M}(\mathbb{R}) \rightarrow \mathcal{M}(\mathbb{R})
$$

we don't know how to add a $2 \times 2$ matrix with a $3 \times 3$ one.

## Closed set

## Definition 4.2

Let $S$ be a set and $H$ a subset of $S$. $H$ is said to be closed with respect to the operation $*$ defined in $S$ iff

$$
\forall a, b \in H \quad a * b \in H
$$

Then we may define the binary operation in H :

$$
\begin{aligned}
*: H \times H & \rightarrow H \\
*(a, b) & =a * b
\end{aligned}
$$

which is called the binary operation induced in $H$.

## Closed set

## Example

Let $S=\mathbb{Z}$ and $H=\left\{n^{2} \mid n \in \mathbb{Z}^{+}\right\}=\{1,4,9,16,25,36, \ldots\} . H$ is not closed with respect to addition. For example:

$$
\begin{aligned}
& 1 \in H \\
& 4 \in H
\end{aligned} \text { but } 1+4 \notin H
$$

## Example

Let $S=\mathbb{Z}$ and $H=\left\{n^{2} \mid n \in \mathbb{Z}^{+}\right\}=\{1,4,9,16,25,36, \ldots\} . H$ is closed with respect to multiplication. For example:

$$
\begin{aligned}
& n^{2} \in H \\
& m^{2} \in H
\end{aligned} \text { and } n^{2} \cdot m^{2}=(n m)^{2} \in H
$$

## Closed set

## Example

Let $S$ be the set of real-valued functions with a single real argument $S=\{\mathbb{R} \rightarrow \mathbb{R}\}$. Let us define the addition of functions as

$$
\begin{aligned}
+:(\mathbb{R} \rightarrow \mathbb{R}) \times(\mathbb{R} \rightarrow \mathbb{R}) & \rightarrow \mathbb{R} \rightarrow \mathbb{R} \\
(f+g)(x) & =f(x)+g(x)
\end{aligned}
$$

Similarly for the multiplication and subtraction of functions. Let us define the composition of functions as

$$
\begin{array}{rll}
\circ:(\mathbb{R} \rightarrow \mathbb{R}) \times(\mathbb{R} \rightarrow \mathbb{R}) & \rightarrow \mathbb{R} \rightarrow \mathbb{R} \\
(f \circ g)(x) & =f(g(x))
\end{array}
$$

$S$ is closed with respect to addition, subtraction, multiplication and composition.

## Definition of a binary operation

## Example

To define a binary operation either we give the full table (intensional definition) as in
or we give a rule to compute it (extensional definition) as in

$$
a * b=(a+b) \bmod 3
$$

## Properties of a binary operation

## Definition 4.3 (Commutativity)

A binary operation is commutative iff

$$
a * b=b * a
$$

## Example

* is commutative because its definition table is symmetric with respect to the main diagonal, but $\triangle$ is not commutative.


## Properties of a binary operation

## Definition 4.4 (Associativity)

A binary operation is associative iff

$$
(a * b) * c=a *(b * c)
$$

## Example

$\triangle$ is not associative because

$$
\begin{aligned}
& (0 \triangle 0) \triangle 0=1 \triangle 0=1 \\
& 0 \triangle(0 \triangle 0)=0 \triangle 1=2
\end{aligned}
$$

But * is associative

$$
\begin{aligned}
& (0 * 0) * 0=0 * 0=0 \\
& 0 *(0 * 0)=0 * 0=0
\end{aligned}
$$

We would have to test all possible triples, but after a a little bit of work we could show that $*$ is associative.

## Properties of a binary operation

## Example

Function composition is associative although not commutative.
Proof
Function composition is not commutative

$$
(f \circ g)(x)=f(g(x)) \neq g(f(x))=(g \circ f)(x)
$$

Function composition is associative

$$
((f \circ g) \circ h)(x)=(f \circ g)(h(x))=f(g(h(x)))=f((g \circ h)(x))=(f \circ(g \circ h))(x)
$$

## Properties of a binary operation

## Example

A function may not be well defined. For instance,

$$
\begin{aligned}
1: \mathbb{Q} \times \mathbb{Q}) & \rightarrow \mathbb{Q} \\
a / b & =\frac{a}{b}
\end{aligned}
$$

is not well defined for $b=0 \in \mathbb{Q}$

## Example

A function may not be closed in $S$. For instance,

$$
\begin{array}{rll}
/: \mathbb{Z} \times \mathbb{Z}) & \rightarrow \mathbb{Z} \\
a / b & = & \frac{a}{b}
\end{array}
$$

is not closed because $a=1 \in \mathbb{Z}, b=3 \in \mathbb{Z}$ but $\frac{1}{3} \notin \mathbb{Z}$.

## Properties of a binary operation

## Definition 4.5 (Existence of a neutral element)

A binary operation has a neutral element, $e$, iff

$$
\forall a \in S \quad a * e=e * a=a
$$

## Example

0 is the neutral element of addition in $\mathbb{R}$ because

$$
\forall r \in \mathbb{R} \quad r+0=0+r=r
$$

1 is the neutral element of multiplication in $\mathbb{R}$ because

$$
\forall r \in \mathbb{R} \quad r \cdot 1=1 \cdot r=r
$$

Addition in $\mathbb{N}$ has no neutral element since $0 \notin \mathbb{N}$.

## Properties of a binary operation

## Definition 4.6 (Existence of an inverse element)

A binary operation has an inverse element iff

$$
\forall a \in S \quad \exists b \in S \mid a * b=b * a=e
$$

being $e$ the neutral element of $*$.

## Example

The inverse element of 2 with respect to addition in $\mathbb{R}$ is -2 because

$$
2+(-2)=(-2)+2=0
$$

The inverse element of 2 with respect to multiplication in $\mathbb{R}$ is $\frac{1}{2}$ because

$$
2 \cdot \frac{1}{2}=\frac{1}{2} \cdot 2=1
$$

Multiplication in $\mathbb{N}$ has no inverse element since $\forall n \in \mathbb{N} \quad \frac{1}{n} \notin \mathbb{N}$.

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## Groups and subgroups

## Introduction

Groups and subgroups are algebraic structures. They are the ones that allow solving equations like

$$
x+x=a \Rightarrow x=\frac{a}{2}
$$

and that the equation

$$
x \cdot x=a
$$

does not have a solution in $\mathbb{R}$ if $a<0$.
We'll see that defining a group amounts to define the elements belonging to the group as well as the operations that can be used with them.

## Groups

## Definition 5.1 (Group)

Given a set $S$ and a binary operation * defined on $S$, the pair $(S, *)$ is a group if $G$ is closed under * and

$$
\text { G1. } * \text { is associative in } S
$$

G2. * has a neutral element in $S$
G3. * has an inverse element in $S$

## Definition 5.2 (Abelian group)

$(S, *)$ is an abelian group if $(S, *)$ is a group and $*$ is commutative.

## Definition 5.3 (Subgroup)

Let $(S, *)$ be a group. Let $H$ be a subset of $S, H \subseteq S$, and $*_{H}$ be the $*$ induced operation in $H$. The pair $\left(H, *_{H}\right)$ is a subgroup of $(S, *)$ if it verifies the conditions to be a group.

## Groups

## Example

Consider $S=\left\{z \in \mathbb{C} \mid z=e^{i \varphi} \quad \forall \varphi \in \mathbb{R}\right\} .(U, \cdot)$ is a group.


## Proof

G1. . is associative in $S$

$$
\begin{gathered}
z_{1}\left(z_{2} z_{3}\right)=e^{i \varphi_{1}}\left(e^{i \varphi_{2}} e^{i \varphi_{3}}\right)=e^{i \varphi_{1}}\left(e^{i\left(\varphi_{2}+\varphi_{3}\right)}\right)=e^{i\left(\varphi_{1}+\varphi_{2}+\varphi_{3}\right)} \\
\left(z_{1} z_{2}\right) z_{3}=\left(e^{i \varphi_{1}} e^{i \varphi_{2}}\right) e^{i \varphi_{3}}=\left(e^{i \varphi_{1}+\varphi_{2}}\right) e^{i \varphi_{3}}=e^{i\left(\varphi_{1}+\varphi_{2}+\varphi_{3}\right)}
\end{gathered}
$$

## Groups

## Example (continued)

## Proof

G2. . has a neutral element in $S$
$1=e^{i 0} \in S$

$$
\begin{aligned}
& z \cdot 1=e^{i \varphi} e^{i 0}=e^{i(\varphi+0)}=e^{i \varphi}=z \\
& 1 \cdot z=e^{i 0} e^{i \varphi}=e^{i(0+\varphi)}=e^{i \varphi}=z
\end{aligned}
$$

G3. . has an inverse element in $S$
For each $z=e^{i \varphi}$, its inverse element with respect to - is $z^{-1}=e^{-i \varphi}$

$$
\begin{gathered}
z z^{-1}=e^{i \varphi} e^{-i \varphi}=e^{i(\varphi-\varphi)}=e^{i 0}=1 \\
z^{-1} z=e^{-i \varphi} e^{i \varphi}=e^{i(-\varphi+\varphi)}=e^{i 0}=1
\end{gathered}
$$

## Groups

## Example

- $(\mathbb{N},+)$ is not a group because it has no neutral element.
- ( $\mathbb{N} \cup\{0\},+$ ) is not a group because it has no inverse element.
- $(\mathbb{Z},+),(\mathbb{Q},+),(\mathbb{R},+),(\mathbb{C},+)$ and $\left(\mathbb{R}^{n},+\right)$ are abelian groups.
- $\left(\mathcal{M}_{m \times n},+\right)$ is an abelian group.
- ( $\mathbb{R}, \cdot)$ is not a group because 0 has no inverse.
- $\left.\left(\mathcal{M}_{n \times n}(\mathbb{R})\right), \cdot\right)$ is not a group because $\left(\begin{array}{cccc}0 & 0 & \ldots & 0 \\ 0 & 0 & \ldots & 0 \\ \ldots & \ldots & \ldots & \ldots \\ 0 & 0 & \ldots & 0\end{array}\right)$ has no inverse.
- Let $S \in \mathcal{M}_{n \times n}(\mathbb{R})$ be the set of invertible matrices of size $n \times n$. $(S, \cdot)$ is a group (although not abelian). It is called the General Linear Group of degree $n(G L(n, \mathbb{R}))$.


## Groups

## Example

The existence of groups is what allows us to solve equations. For instance, consider the equation

$$
5+x=2
$$

and its solution in the group $(\mathbb{Z},+)$

$$
\begin{array}{rlrl}
5+x & =2 & & \text { [Addition of the inverse of } 5 \text { with respect to }+ \text { in both } \\
-5+(5+x) & =-5+2 \quad & \text { [Addition is associative ] } \\
(-5+5)+x & =-3 & \text { [Definition of inverse] } \\
0+x & =-3 & \text { [Definition of neutral element] }
\end{array}
$$

## Groups

## Example

Consider the equation

$$
2 x=3
$$

and its solution in the group $(\mathbb{Q}, \cdot)$

$$
\begin{aligned}
2 x & =3 \quad \text { [Multiplication by the inverse of } 2 \text { in both sides] } \\
\frac{1}{2}(2 x) & =\frac{1}{2} 3 \quad \text { [Multiplication is associative ] } \\
\left(\frac{1}{2} 2\right) x & =\frac{2}{3} \quad \text { [Definition of inverse] } \\
1 x & =\frac{2}{3} \quad \text { [Definition of neutral element] }
\end{aligned}
$$

## Groups

## Theorem 5.1 (Cancellation laws)

Given any group $(S, *), \forall a, b, c \in S$ it is verified

- Left cancellation: $a * b=a * c \Rightarrow b=c$
- Right cancellation: $b * a=c * a \Rightarrow b=c$

Theorem 5.2 (Existance of a unique solution of linear equations)
Given any group $(S, *), \forall a, b \in S$ the linear equations

$$
a * x=b \text { and } y * a=b
$$

always have a unique solution in $S$.

Theorem 5.3 (Properties of the inverse)
Given any group $(S, *), \forall a \in S$ its inverse is unique and $\forall a, b \in S$

$$
(a * b)^{-1}=\left(b^{-1}\right) *\left(a^{-1}\right)
$$

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## Homomorphisms

## Example

Consider the sets $S=\{a, b, c\}$ and $S^{\prime}=\{A, B, C\}$ with the operations $*: S \times S \rightarrow S$ and $*^{\prime}: S^{\prime} \times S^{\prime} \rightarrow S^{\prime}$

$$
\begin{array}{c|cccc|ccc}
x * y & y=a & y=b & y=c \\
\hline x=a & a & b & c \\
x=b & b & c & a & & \text { and } & & x *^{\prime} y \\
x=A & y=A & y=B & y=C \\
x=c & c & a & b & & x=C & B & C \\
x=B & B & C & A \\
x=C & & & A & B
\end{array}
$$

We may construct a function that "translates" elements in $S$ into elements in $S^{\prime}$ with the "same properties".

$$
\begin{aligned}
\phi: S & \rightarrow S^{\prime} \\
\phi(a) & = \\
\phi(b) & =B \\
\phi(c) & =C
\end{aligned}
$$

We note that

$$
b * c=a \Rightarrow \phi(b) *^{\prime} \phi(c)=\phi(a) \Rightarrow B *^{\prime} C=A
$$

## Homomorphisms

## Definition 6.1 (Group homomorphism)

Given two groups $(S, *)$ and $\left(S^{\prime}, *^{\prime}\right)$, the function $\phi: S \rightarrow S^{\prime}$ is a group homomorphism iff $\forall a, b \in S$

$$
\phi(a * b)=\phi(a) *^{\prime} \phi(b)
$$

## Definition 6.2 (Group isomorphism)

Given two groups $(S, *)$ and $\left(S^{\prime}, *^{\prime}\right)$, the function $\phi: S \rightarrow S^{\prime}$ is a group isomorphism iff it is a group homomorphism and it is bijective.

## Homomorphisms

## Example

Consider the two groups $\left(\mathbb{R}^{n},+\right)$ and $\left(\mathbb{R}^{m},+\right)$ and a matrix $A \in \mathcal{M}_{m \times n}(\mathbb{R})$. The application

$$
\begin{aligned}
\phi: \mathbb{R}^{n} & \rightarrow \mathbb{R}^{m} \\
\phi(\mathbf{x}) & =A \mathbf{x}
\end{aligned}
$$

is a group homomorphism because

$$
\phi(\mathbf{u}+\mathbf{v})=A(\mathbf{u}+\mathbf{v})=A \mathbf{u}+A \mathbf{v}=\phi(\mathbf{u})+\phi(\mathbf{v})
$$

## Example

Consider the two groups $(G L(n, \mathbb{R}), \cdot)$ and $(\mathbb{R}, \cdot)$. The application

$$
\begin{aligned}
\phi: G L(n, \mathbb{R}) & \rightarrow \mathbb{R} \\
\phi(A) & =\operatorname{det}\{A\}
\end{aligned}
$$

is a group homomorphism because

$$
\phi(A B)=\operatorname{det}\{A B\}=\operatorname{det}\{A\} \operatorname{det}\{B\}=\phi(A) \cdot \phi(B)
$$

## Homomorphisms

## Theorem 6.1

Let $\phi: S \rightarrow S^{\prime}$ be a group homomogrphism between two groups. Then,

- $\phi(e)=e^{\prime}$
- $\phi\left(a^{-1}\right)=(\phi(a))^{-1}$


## Definition 6.3 (Kernel of a group homomorphism)

Let $\phi: S \rightarrow S^{\prime}$ be a group homomogrphism between two groups. Then, the kernel of $\phi$ is the set

$$
\operatorname{Ker}\{\phi\}=\left\{x \in S \mid \phi(x)=e^{\prime}\right\}
$$

## Example

Let $\phi(\mathbf{x})=A \mathbf{x}$. Then,

$$
\operatorname{Ker}\{\phi\}=\left\{x \in \mathbb{R}^{n} \mid A \mathbf{x}=\mathbf{0}\right\}=\operatorname{Nul}\{A\}
$$

## Isomorphisms

## Theorem 6.2 (Isomorphisms and cardinality)

If two groups $(S, *)$ and $\left(S^{\prime}, *^{\prime}\right)$ are isomorph (i.e., there exists an isomorphism between the two groups), then $S$ and $S^{\prime}$ have the same cardinality.


## Isomorphisms

## Example

- $\mathbb{Q}$ and $\mathbb{R}$ cannot be isomorph because the cardinality of $\mathbb{Q}$ is $\aleph_{0}$ and the cardinality of $\mathbb{R}$ is $\aleph_{1}$.
- There are as many natural numbers as natural even numbers. In other words, the cardinality of $\mathbb{N}$ and $2 \mathbb{N}$ are the same. The reason is that the function $\phi(n)=2 n$ is an isomorphism between $\mathbb{N}$ and $2 \mathbb{N}$.


## Example

Consider the set $\mathbb{R}_{c}=[0, c) \in \mathbb{R}$ and the operation $x+{ }_{c} y=(x+y) \bmod c$. The pair $\left(\mathbb{R}_{c},+_{c}\right)$ is a group. Consider now the two particular cases $\left(\mathbb{R}_{2 \pi},+2 \pi\right)$ and ( $\mathbb{R}_{1},+{ }_{1}$ ) and the mapping

$$
\begin{aligned}
\phi: \mathbb{R}_{2 \pi} & \rightarrow \mathbb{R}_{1} \\
\phi(x) & =\frac{x}{2 \pi}
\end{aligned}
$$

$\phi$ is an isomorphism between $\left(\mathbb{R}_{2 \pi},+_{2 \pi}\right)$ and $\left(\mathbb{R}_{1},+_{1}\right)$. In fact, all $\left(\mathbb{R}_{c},+_{c}\right)$ groups are isomorph to any other $\left(\mathbb{R}_{c^{\prime}},+{ }_{c^{\prime}}\right)$ group.

## Isomorphisms

Cardinality is a group property. The nice things about isomorphisms is that they preserve group properties.

## Theorem 6.3

If two groups $(S, *)$ and $\left(S^{\prime}, *^{\prime}\right)$ are isomorph, then

- If $*$ is commutative, so is $*^{\prime}$.
- If there is an order relation in S, it can be "translated" into an order relation in $S^{\prime}$.
- If $\forall s \in S$ there exists a solution in $S$ of the equation $x * x=s$, then $\forall s^{\prime} \in S^{\prime}$ there exists a solution in $S^{\prime}$ of the equation $x *^{\prime} x=s^{\prime}$.
- If $\forall a, b \in S$ there exists a solution in $S$ of the equation $a * x=b$, then $\forall a^{\prime}, b^{\prime} \in S^{\prime}$ there exists a solution in $S^{\prime}$ of the equation $a^{\prime} *^{\prime} x=b^{\prime}$.
- The kernel of any isomorphism $\phi$ between $(S, *)$ and $\left(S^{\prime}, *^{\prime}\right)$ is $\operatorname{Ker}\{\phi\}=\{e\}$ being $e$ the neutral element of $*$ in $S$.


## Isomorphisms

## Example

$((Z),+)$ is not isomorph to $((Q),+)$ because the equation

$$
x+x=s
$$

has a solution in $\mathbb{Q}$ for any $s \in \mathbb{Q}$ (that is $x=\frac{s}{2}$ ), but it does not have a solution in $\mathbb{Z}$ for any $s \in \mathbb{Z}$ (it only has a solution in $\mathbb{Z}$ if $s$ is an even number).

## Example

$((R), \cdot)$ is not isomorph to $((C), \cdot)$ because the equation

$$
x \cdot x=z
$$

has two solution in $\mathbb{C}$ for any $z \in \mathbb{C}$ (in fact there are two solutions, if $z=r e^{i \theta}$, then $x= \pm r e^{i \frac{\theta}{2}}$ are the two solutions), but it does not have a solution in $\mathbb{R}$ for any $z \in \mathbb{R}$ (it only has a solution in $\mathbb{R}$ if $z$ is a non-negative number).

## Outline

(10) Abstract algebra

- Sets
- Relations and functions
- Partitions and equivalence relationships
- Binary operations
- Groups and subgroups
- Homomorphisms and isomorphisms
- Algebraic structures


## Algebraic structures

## Algebraic structures

Algebraic structures are tools that help us to define operate on numbers and elements within a set, solve equations, etc.

Set $\mathbf{S}$ with binary operation +
Operation + is associative

| monoid | Existence of identity element of $\boldsymbol{+}$ in $\boldsymbol{S} \triangle$ |
| :---: | :---: |
| group | Existence of inverse elements of + in $\boldsymbol{S}_{\triangle}$ |
| abelian group | Commutativity of + |
|  | Associative binary operation - |
| pseudo-ring | Distributivity of - over + |
| ring | Existence of identity element of $\operatorname{in} \boldsymbol{S} \triangle$ |
| commutative ring | Commutativity of • |
| field | Existence of inverse elements of $\bullet$ in $\boldsymbol{S} \triangle$ |

## Algebraic structures

## Definition 7.1 (Ring)

The tuple $(S, *, \circ)$ is a ring iff

$$
\text { R1. }(S, *) \text { is an abelian group. }
$$

$R 2$. ○ is associative.
$R 3$. o is distributive with respect to $*$, i.e., $\forall a, b, c \in S$

- Left-distributive: $a \circ(b * c)=(a \circ b) *(a \circ c)$
- Right-distributive: $(a * b) \circ c=(a \circ c) *(b \circ c)$


## Example

- $(\mathbb{Z},+, \cdot),(\mathbb{Q},+, \cdot),(\mathbb{R},+, \cdot),(\mathbb{C},+, \cdot)$ are rings.
- $\left(\mathcal{M}_{m \times n}(\mathbb{R}),+, \cdot\right)$ is a ring.
- $(\mathbb{R} \rightarrow \mathbb{R},+, \cdot)$ is a ring.


## Algebraic structures

## Theorem 7.1 (Properties of rings)

Let $(S, *, \circ)$ be a ring and let e be the neutral element of $*$ in $S$. For any $a \in S$, let $a^{\prime}$ be the inverse of a with respect to the operation $*$. Then $\forall a, b \in S$

- $a \circ e=e \circ a=e$.
- $a \circ b^{\prime}=a^{\prime} \circ b=(a \circ b)^{\prime}$
- $a^{\prime} \circ b^{\prime}=a \circ b$


## Example

Consider the ring $(\mathbb{R},+, \cdot)$. We are used to the properties $\forall a, b \in \mathbb{R}$

- $a \cdot 0=0 \cdot a=0$.
- $a \cdot(-b)=(-a) \cdot b=-(a \cdot b)$
- $(-a) \cdot(-b)=a \cdot b$

But, as stated by the previous theorem, these are properties of all rings.

## Algebraic structures

## Definition 7.2 (Kinds of rings)

A ring $(S, *, \circ)$ is

- commutative iff $\circ$ is commutative.
- unitary iffo has a neutral element (referred as 1 ).
- divisive if it is unitary and

$$
\forall a \in S-\{e\} \quad \exists!a^{-1} \in S, \mid a \circ a^{-1}=a^{-1} \circ a=1
$$

That is each element has a multiplicative inverse.

## Example

- $(\mathbb{P},+, \cdot)$ the set of polynomials with coefficients from a ring is a ring.


## Algebraic structures

## Definition 7.3 (Field (cuerpo))

A divisive, commutative ring is called a field.

## Example

- $(\mathbb{Q},+, \cdot),(\mathbb{R},+, \cdot)$, and $(\mathbb{C},+, \cdot)$ are fields.
- $(\mathbb{Z},+, \cdot)$ is not a field because multiplication has not an inverse in $\mathbb{Z}$.


## Algebraic structures

## Definition 7.4 (Vector space over a field)

Consider a field $(\mathbb{K}, *, \circ)$. A vector space over this field is a tuple $(V,+, \cdot)$ so that $V$ is a set whose elements are called vectors, and $+: V \times V \rightarrow V$ is a binary operation under which $V$ is closed, $\cdot: \mathbb{K} \times V \rightarrow V$ is an operation between scalars in the field $(\mathbb{K})$ and vectors in the vector space ( $V$ ) such that $\forall a, b \in \mathbb{K}, \forall \mathbf{u}, \mathbf{v} \in V$

V1. $(V,+)$ is an abelian group.
V2. $(a \cdot \mathbf{u}) \in V$
V3. $a \cdot(b \cdot \mathbf{u})=(a \circ b) \cdot \mathbf{u}$
V4. $(a * b) \cdot \mathbf{u}=a \cdot \mathbf{u}+b \cdot \mathbf{u}$
V5. $a \cdot(\mathbf{u}+\mathbf{v})=a \cdot \mathbf{u}+a \cdot \mathbf{v}$
V6. $1 \cdot \mathbf{u}=\mathbf{u}$

## Algebraic structures

## Examples

- $\left(\mathbb{R}^{n},+, \cdot\right)$ and $\left(\mathbb{C}^{n},+, \cdot\right)$.
- $\left(\mathcal{M}_{m \times n}(\mathbb{R}),+, \cdot\right)$ : the set of matrices of a given size with coefficients in a field.
- $(\mathbb{P},+, \cdot)$ : the set of polynomials with coefficients in a field.
- $(\{X \rightarrow V\},+, \cdot)$ : the set of all functions from an arbitrary set $X$ onto an arbitrary vector space $V$.
- The set of all continuous functions is a vector space.
- The set of all linear maps between two vector spaces is also a vector space.
- The set of all infinite sequences of values from a field is also a vector space.


## Algebraic structures

## Definition 7.5 (Algebra)

Consider a vector space $(V,+, \cdot)$ over a field $(\mathbb{K}, *, \circ)$ and a binary operation $\bullet: V \times V \rightarrow V .(V,+, \cdot, \bullet)$ is an algebra iff $\forall a, b \in \mathbb{K}, \forall \mathbf{u}, \mathbf{v}, \mathbf{w} \in V$ A1. Left distributivity: $(\mathbf{u}+\mathbf{v}) \bullet \mathbf{w}=\mathbf{u} \bullet \mathbf{w}+\mathbf{v} \bullet \mathbf{w}$ A2. Right distributivity: $\mathbf{u} \bullet(\mathbf{v}+\mathbf{w})=\mathbf{u} \bullet \mathbf{v}+\mathbf{u} \bullet \mathbf{w}$ A3. Compatibility with scalars: $(a \cdot \mathbf{u}) \bullet(b \cdot \mathbf{v})=(a \circ b) \cdot(\mathbf{u} \bullet \mathbf{v})$

## Examples

- Real numbers ( $\mathbb{R}$ ) are an algebra ("1D").
- Complex numbers $(\mathbb{C})$ are an algebra ("2D").
- Quaternions are an algebra ("4D").


## Outline

(10) Abstract algebra

- Sets
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- Homomorphisms and isomorphisms
- Algebraic structures


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