# Mathematical modelling 

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Faculty of Computer and Information Science University of Ljubljana

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An application of SVD: principal component analysis or PCA
PCA is a very well-known and efficient method for data compression, dimension reduction, ...

Due to its importance in different fields, it has many other names: discrete
Karhunen-Loève transform (KLT), Hotelling transform, empirical orthogonal functions (EOF), ...

Let $\left\{X_{1}, \ldots, X_{m}\right\}$ be a sample of vectors from $\mathbb{R}^{n}$.
In applications, often $m \ll n$, where $n$ is very large, for example, $X_{1}, \ldots, X_{m}$ can be

- vectors of gene expressions in $m$ tissue samples or
- vectors of grayscale in images
- bag of words vectors, with components corresponding to the numbers of certain words from some dictionary in specific texts, ...,
or $n \ll m$ for example if the data represents a point cloud in a low dimensional space $\mathbb{R}^{n}$ (for example in the plane).

We will assume that $m \ll n$. Also assume that the data is centralized, i.e., the centeroid is in the origin

$$
\mu=\frac{1}{m} \sum_{i=1}^{m} X_{i}=0 \in \mathbb{R}^{n}
$$

If not, we substract $\mu$ from all vectors in the data set.
A matrix norm $\|\cdot\|: \mathbb{R}^{n \times m} \rightarrow \mathbb{R}$ is a function, which generalizes the notion of the absolute value for numbers to matrices. It is used to measure a distance between matrices. In contrast with the absolute value, which is unique up to multiplication with a positive constant, there are many different matrix norms.

Two important matrix norms are the following:

1. Spectral norm $\|\cdot\|_{2}$ :

$$
\|A\|_{2}:=\max _{\|x\|_{2}=1}\|A x\|_{2}=\max _{j=1, \ldots, \min (n, m)} \sigma_{j}(A)
$$

2. Frobenius norm $\|\cdot\|_{F}$ :

$$
\|A\|_{F}:=\sqrt{\sum_{i, j} a_{i, j}^{2}}=\sqrt{\sum_{j=1, \ldots, \min (n, m)} \sigma_{j}(A)^{2}}
$$

Let

$$
X=\left[\begin{array}{llll}
X_{1} & X_{2} & \cdots & X_{m}
\end{array}\right]^{T}
$$

be the matrix of dimension $m \times n$ with data in the rows.
Let $X^{T} X \in \mathbb{R}^{m \times m}$ and $X X^{T} \in \mathbb{R}^{n \times n}$ be the covariance matrices of the data.

- The principal values of the data set $\left\{X_{1}, \ldots, X_{r}\right\}$ are the nonzero eigenvalues $\lambda_{i}=\sigma_{i}^{2}$ of the covariance matrices (where $\sigma_{i}$ are the singular values of $X$ ).
- The principal directions in $\mathbb{R}^{n}$ are corresponding eigenvectors $v_{1}, \ldots, v_{r}$, i.e. the columns of the matrix $V$ from the SVD of $X$. The remaining clolumns of $V$ (i.e. the eigenvectors correspondong to 0 ) form a basis of the null space of $X$.
- The first column $v_{1}$, the first principal direction, corresponds to the direction in $\mathbb{R}^{n}$ with the largest variance in the data $X_{i}$, that is, the most informative direction for the data set, the second the second most important, ...
- The principal directions in $\mathbb{R}^{m}$ are the columns $u_{1}, \ldots, u_{r}$ of the matrix $U$ and represent the coefficients in the linear decomposition of the vectors $X_{1}, \ldots, X_{m}$ along the orthonormal basis $v_{1}, \ldots v_{n}$ of $\mathbb{R}^{n}$.

PCA provides a linear dimension reduction method based on a projection of the data from the space $\mathbb{R}^{n}$ into a lower dimensional subspace spanned by the first few principal vectors $v_{1}, \ldots, v_{k}$ in $\mathbb{R}^{n}$.

The idea is to approximate

$$
X_{i}=\sigma_{1} u_{1, i} v_{1}+\cdots+\sigma_{m} u_{m, i} v_{m} \cong \sigma_{1} u_{1, i} v_{1}+\cdots+\sigma_{k} u_{k, i} v_{k}
$$

with the first $k$ most informative directions in $\mathbb{R}^{n}$ and supress the last $m-k$.

PCA has the following amazing property:

## Theorem

Among all possible projections of $p: \mathbb{R}^{n} \rightarrow \mathbb{R}^{k}$ onto a $k$-dimensional subspace, PCA provides the best in the sense that the errors

$$
\|X-p(X)\|_{F}^{2} \quad \text { and } \quad\|X-p(X)\|_{2}^{2}
$$

where $p(X)=\left[\begin{array}{lll}p\left(X_{1}\right) & \cdots & p\left(X_{m}\right)\end{array}\right]^{T}$, are the smallest possible.

## Chapter 3:

## Nonlinear models

- Definition and examples
- Systems of nonlinear equations
- Vector functions of vector variables
- Derivative and Jacobian matrix
- Linear approximation
- Newton's method for square systems
- Univariate case: Tangent method
- Use in optimization
- Gauss-Newton's method for rectangular systems


## 3. Nonlinear models

General formulation
Given is a sample of points $\left\{\left(x_{1}, y_{1}\right), \ldots,\left(x_{m}, y_{m}\right)\right\}, x_{i} \in \mathbb{R}^{n}, y_{i} \in \mathbb{R}$.
The mathematical model is nonlinear if the function

$$
\begin{equation*}
y=F\left(x, a_{1}, \ldots, a_{p}\right) \tag{1}
\end{equation*}
$$

is a nonlinear function of the parameters $a_{i}$. This means it cannot be written in the form

$$
y=a_{1} f_{1}(x)+a_{2} f_{2}(x)+\ldots+a_{p} f_{p}(x)
$$

where each $f_{i}: \mathbb{R}^{n} \rightarrow \mathbb{R}$ is some function.
Plugging each data points into (1) we obtain a system of nonlinear equations

$$
\begin{align*}
y_{1} & =F\left(x_{1}, a_{1}, \ldots, a_{p}\right), \\
& \vdots  \tag{2}\\
y_{m} & =F\left(x_{m}, a_{1}, \ldots, a_{p}\right),
\end{align*}
$$

in the parameters $a_{1}, \ldots, a_{p} \in \mathbb{R}$.

1. Exponential decay or growth: $F(x, a, k)=a e^{k x}, a$ and $k$ are parameters.

A quantity $y$ changes at a rate proportional to its current value, which can be described by the differential equation

$$
\frac{d y}{d x}=k y
$$

The solution to this equation (obtained by the use of separation of variables) is $y=F(x, a, k)$.

2. Gaussian model: $F(x, a, b, c)=a e^{-\left(\frac{x-b}{c}\right)^{2}}, a, b, c \in \mathbb{R}$ parameters. $a$ is the value of the maximum obtained at $x=b$ and $c$ determines the width of the curve.

It is used in statistics to describe the normal distribution, but also in signal and image processing.
In statistics $a=\frac{1}{\sigma \sqrt{2 \pi}}, b=\mu, c=\sqrt{2 \sigma}$, where $\mu, \sigma$ are the expected value and the standard deviation of a normally distributed random variable.

3. Logistic model: $F(x, a, b, k)=\frac{a}{\left(1+b e^{-k x}\right)}, k>0$

The logistic function was devised as a model of population size by adjusting the exponential model which also considers the saturation of the environment, hence the growth first changes to linear and then stops.

The logistic function $F(x, a, b, k)$ is a solution of the first order non-linear differential equation

$$
\frac{d y(x)}{d x}=k y(x)\left(1-\frac{y(x)}{a}\right)
$$


4. In the area around a radiotelescope the use of microwave ovens is forbidden, since the radiation interferes with the telescope. We are looking for the location $(a, b)$ of a microwave oven that is causing problems.

The radiation intensity decreases with the distance $r$ from the source according to $u(r)=\frac{\alpha}{1+r}$. In cartesian coordinates:

$$
u(x, y)=\frac{\alpha}{1+\sqrt{(x-a)^{2}+(y-b)^{2}}}
$$

where $(a, b)$ is a position of the microwave.
Task: Find the position of the microwave, if the measured values of the signal at three locations are $u(0,0)=0.27, u(1,1)=0.36$ in $u(0,2)=0.3$.

This gives the following system of equations for the parameters $\alpha, a, b$ :

$$
\begin{aligned}
\frac{\alpha}{1+\sqrt{a^{2}+b^{2}}} & =0.27 \\
\frac{\alpha}{1+\sqrt{(1-a)^{2}+(1-b)^{2}}} & =0.36 \\
\frac{\alpha}{1+\sqrt{a^{2}+(2-b)^{2}}} & =0.3
\end{aligned}
$$

- Our goal is to fit the data points

$$
\left\{\left(x_{1}, y_{1}\right), \ldots,\left(x_{m}, y_{m}\right)\right\}, \quad x_{i} \in \mathbb{R}^{n}, y_{i} \in \mathbb{R}
$$

- We choose a fitting function

$$
F\left(x, a_{1}, \ldots, a_{p}\right)
$$

which depends on the unknown parameters $a_{1}, \ldots, a_{p}$.

- Equivalent formulation of the system (2) (which will be more suitable for solving with numerical algorithms) is:

1. For $i=1, \ldots, m$ define the functions

$$
g_{i}: \mathbb{R}^{p} \rightarrow \mathbb{R} \quad \text { by the rule } \quad g_{i}\left(a_{1}, \ldots, a_{p}\right)=y_{i}-F\left(x_{i}, a_{1}, \ldots, a_{p}\right) .
$$

2. Solve or approximate the following system by the least squares method

$$
\begin{gather*}
g_{1}\left(a_{1}, \ldots, a_{p}\right)=0 \\
\vdots  \tag{3}\\
g_{m}\left(a_{1}, \ldots, a_{p}\right)=0
\end{gather*}
$$

In a compact way (3) can be expressed by introducing a vector function

$$
\begin{equation*}
G: \mathbb{R}^{p} \rightarrow \mathbb{R}^{m}, \quad G\left(a_{1}, \ldots, a_{p}\right)=\left(g_{1}\left(a_{1}, \ldots, a_{p}\right), \ldots, g_{m}\left(a_{1} \ldots, a_{p}\right)\right), \tag{4}
\end{equation*}
$$

and search for the tuples $\left(a_{1}, \ldots, a_{p}\right)$ that solve the system (or minimize the norm of the left-hand side)

$$
\begin{equation*}
G\left(a_{1}, \ldots, a_{p}\right)=(0, \ldots, 0) \tag{5}
\end{equation*}
$$

## Remark

Solving (5) is a difficult problem. Even if the exact solution exists, it is not easy (or even impossible) to compute. For example, there does not even exist an analytic formula to determine roots of a general polynomial of degree 5 or more.

### 3.1 Vector functions of a vector variable

Neccessary terminology to achieve our plan
$G$ from (4) is an example of

- a vector function: since it maps into $\mathbb{R}^{m}$, where $m$ might be bigger than 1.
- a vector variable: since it maps from $\mathbb{R}^{p}$, where $p$ might be bigger than 1.


## Remark

- If $m=1$ and $p>1$, then $G$ is a usual multivariate function.
- If $m=1$ and $p=1$, then $G$ is a usual (univariate) function.

For easier reference in the continuation we call $g_{1}, \ldots, g_{m}$ from (4) the component (or coordinate) functions of $G$.

1. A linear vector function $G: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ is such that all the component functions $g_{i}$ are linear:

$$
g_{i}\left(x_{1}, \ldots, x_{n}\right)=a_{i 1} \cdot x_{1}+a_{i 2} \cdot x_{2}+\ldots+a_{i n} \cdot x_{n}, \quad \text { where } a_{i j} \in \mathbb{R} . \text { (6) }
$$

In this case

$$
G(x)=A x
$$

where

$$
A=\left[\begin{array}{cccc}
a_{11} & a_{12} & \ldots & a_{1 n} \\
a_{21} & a_{22} & \ldots & a_{2 n} \\
\vdots & \vdots & \ddots & \vdots \\
a_{m 1} & a_{m 2} & \ldots & a_{m n}
\end{array}\right]
$$

2. Adding constants $b_{i} \in \mathbb{R}$ to the left side of (6) we get the definition of an affine linear vector function,

$$
g_{i}\left(x_{1}, \ldots, x_{n}\right)=a_{i 1} x_{1}+a_{i 2} x_{2}+\ldots a_{i n} x_{n}+b_{i}
$$

and then

$$
G(x)=A x+b, \quad \text { where } \quad b=\left[\begin{array}{llll}
b_{1} & b_{2} & \ldots & b_{n}
\end{array}\right]^{T} .
$$

3. Most of the (vector) functions are nonlinear, e.g.,

$$
\begin{aligned}
f: \mathbb{R}^{3} \rightarrow \mathbb{R}^{2}, & f(x, y, z)=\left(x^{2}+y^{2}+z^{2}-1, x+y+z\right), \\
g: \mathbb{R}^{2} \rightarrow \mathbb{R}^{3}, & g(z, w)=\left(z w, \cos z+w^{2}-2, e^{2 z}\right), \\
h: \mathbb{R} \rightarrow \mathbb{R}^{2}, & h(t)=\left(t+3, e^{-3 t}\right) .
\end{aligned}
$$

The derivative of a vector function $F: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ in the point

$$
a:=\left(a_{1}, \ldots, a_{n}\right) \in \mathbb{R}^{n}
$$

is called the Jacobian matrix of $F$ in a:

$$
J_{F}(a)=D F(a)=\left[\begin{array}{ccc}
\frac{\partial f_{1}}{\partial x_{1}}(a) & \cdots & \frac{\partial f_{1}}{\partial x_{n}}(a) \\
\vdots & \ddots & \vdots \\
\frac{\partial f_{m}}{\partial x_{1}}(a) & \cdots & \frac{\partial f_{m}}{\partial x_{n}}(a)
\end{array}\right]
$$

- If $n=m=1$, the $D f(x)=f^{\prime}(x)$ is the usual derivative.


- For general $n$ and $m=1, f$ is a function of $n$ variables and

$$
D f(x)=\operatorname{grad} f(x)
$$

is its gradient.


- For general $m$ and $n, D f(x)=\left[\begin{array}{c}\operatorname{grad} f_{1} \\ \vdots \\ \operatorname{grad} f_{m}\end{array}\right]$ is a vector of gradients of component functions.

1. For an affine linear function $f: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$, given by $f(x)=A x+b$, it is easy to check that

$$
D f(x)=A
$$

2. For a vector function $f: \mathbb{R}^{3} \rightarrow \mathbb{R}^{2}$, given by

$$
f(x, y, z)=\left(x^{2}+y^{2}+z^{2}-1, x+y+z\right)
$$

then

$$
D f(x)=\left[\begin{array}{ccc}
2 x & 2 y & 2 z \\
1 & 1 & 1
\end{array}\right]
$$

A linear approximation of the vector function $f: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}$ at the point $a \in \mathbb{R}^{n}$ is the affine linear function

$$
L_{a}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{m}, \quad L_{a}(x)=A x+b
$$

that satisfies the following conditions:

1. It has the same value as $f$ in $a: L_{a}(a)=f(a)$.
2. It has the same derivative as $f$ at $a: D L_{a}(a)=D f(a)$.

It is easy to check that

$$
L_{a}(x)=f(a)+D f(a)(x-a)
$$

- $n=m=1$ :

$$
L_{a}(x)=f(a)+f^{\prime}(a)(x-a)
$$

The graph $y=L_{a}(x)$ is the tangent to the graph $y=f(x)$ at the point $a$.

- If $n=2$ and $m=1$, then

$$
L_{(a, b)}(x, y)=f(a, b)+\operatorname{grad} f(a, b)\left[\begin{array}{l}
x-a \\
y-b
\end{array}\right]
$$

The graph

$$
z=L_{(a, b)}(x, y)
$$

is the tangent plane to the surface $z=f(x, y)$ at the point $(a, b)$.


The linear approximation of the function

$$
f: \mathbb{R}^{3} \rightarrow \mathbb{R}^{2}, \quad f(x, y, z)=\left(x^{2}+y^{2}+z^{2}-1, x+y+z\right)
$$

at $a=(1,-1,1)$ is the affine linear function

$$
\begin{aligned}
L_{a}(x, y, z) & =f(1,-1,1)+D f(1,-1,1)\left[\begin{array}{l}
x-1 \\
y+1 \\
z-1
\end{array}\right] \\
& =\left[\begin{array}{l}
2 \\
1
\end{array}\right]+\left[\begin{array}{rrr}
2 & -2 & 2 \\
1 & 1 & 1
\end{array}\right]\left[\begin{array}{c}
x-1 \\
y+1 \\
z-1
\end{array}\right] \\
& =\left[\begin{array}{c}
2+2(x-1)-2(y+1)+2(z-1) \\
1+(x-1)+(y+1)+(z-2)
\end{array}\right] \\
& =\left[\begin{array}{rrr}
2 & -2 & 2 \\
1 & 1 & 1
\end{array}\right]\left[\begin{array}{c}
x \\
y \\
z
\end{array}\right]+\left[\begin{array}{c}
-4 \\
0
\end{array}\right] .
\end{aligned}
$$

### 3.2 Solving systems of nonlinear equations

Let $f: D \rightarrow \mathbb{R}^{m}$ be a vector function, defined on some set $D \subset \mathbb{R}^{n}$.
We will study the Gauss-Newton method to solve the system $f(x)=0$ in terms of least squares. This is one of the numerical methods for searching approximate solution of this system. It is based on linear approximations of $f$.

We are searching zeroes of the function $f: D \rightarrow \mathbb{R}, D \subseteq \mathbb{R}$, i.e., we are solving $f(x)=0$.

Newton's or tangent method:
We construct a recursive sequence with:

- $x_{0}$ is an initial term,
- $x_{k+1}$ is a solution of

$$
L_{x_{k}}(x)=f\left(x_{k}\right)+f^{\prime}\left(x_{k}\right)\left(x-x_{k}\right)=0, \text { so } x_{k+1}=x_{k}-\frac{f\left(x_{k}\right)}{f^{\prime}\left(x_{k}\right)} .
$$



## Theorem

The sequence $x_{i}$ converges to a solution $\alpha, f(\alpha)=0$, if:
(1) $0 \neq\left|f^{\prime}(x)\right|$ for all $x \in I$, where I is some interval containing $\alpha$,
(2) $x_{0}$ is sufficiently close to $\alpha$.

Under these assumptions the convergence is quadratic, meaning that:

$$
\text { If we denote by } \varepsilon_{j}=\left|x_{j}-\alpha\right| \text {, then } \varepsilon_{i+1} \leq M \varepsilon_{i}^{2}
$$

where $M$ is some constant. If $f$ is twice differentiable, then

$$
M \leq \max _{x \in I}\left|f^{\prime \prime}(x)\right| / \min _{x \in I}\left|f^{\prime}(x)\right| .
$$

## Proof.

Condition (1) implies in particular that $\alpha$ is a simple zero of $f$. Plugging $\alpha$ in the Taylor expansion of $f$ around $x_{i}$ we get

$$
\begin{align*}
0=f(\alpha) & =f\left(x_{i}\right)+f^{\prime}\left(x_{i}\right)\left(\alpha-x_{i}\right)+\frac{f^{\prime \prime}(\eta)}{2}\left(\alpha-x_{i}\right)^{2} \\
& =f\left(x_{i}\right)+f^{\prime}\left(x_{i}\right)\left(\alpha-x_{i}\right)+\frac{f^{\prime \prime}(\eta)}{2}\left(\alpha-x_{i}\right)^{2} \tag{7}
\end{align*}
$$

where $\eta$ is between $\alpha$ and $x_{i}$. Dividing (7) with $f^{\prime}\left(x_{i}\right)$ we get

$$
0=\frac{f\left(x_{i}\right)}{f^{\prime}\left(x_{i}\right)}-\left(\alpha-x_{i}\right)+\frac{f^{\prime \prime}(\eta)}{2 f^{\prime}\left(x_{i}\right)} e_{i}^{2}
$$

and hence

$$
\left(x_{i}-\frac{f\left(x_{i}\right)}{f^{\prime}\left(x_{i}\right)}\right)-\alpha=x_{i+1}-\alpha=\frac{f^{\prime \prime}(\eta)}{2 f^{\prime}\left(x_{i}\right)} e_{i}^{2} .
$$

Thus,

$$
e_{i+1}=\left|\frac{f^{\prime \prime}(\eta)}{2 f^{\prime}\left(x_{i}\right)}\right| e_{i}^{2}
$$

Now

$$
\left|\frac{f^{\prime \prime}(\eta)}{2 f^{\prime}\left(x_{i}\right)}\right| \leq \frac{\max _{x \in I}\left|f^{\prime \prime}(x)\right|}{\min _{x \in I}\left|f^{\prime}(x)\right|} .
$$

To prove that the sequence converges note that there exists $\delta_{0}>0$ such that

$$
M \delta_{0}<\frac{1}{2}
$$

Hence, if $e_{i} \leq \delta_{0}$, then

$$
e_{i+1}=\left|\frac{f^{\prime \prime}(\eta)}{2 f^{\prime}\left(x_{i}\right)}\right| e_{i}^{2}=\frac{1}{2} e_{i} .
$$

Therefore

$$
\lim _{n \rightarrow \infty} e_{n}=\lim _{n \rightarrow \infty} \frac{1}{2^{n}} \cdot e_{0}=0
$$

Newton's method generalizes to systems of $n$ nonlinear equations in $n$ unknowns:

- $x_{0}$ - initial approximation,
- $x_{k+1}$ - solution of

$$
L_{x_{k}}(x)=f\left(x_{k}\right)+D f\left(x_{k}\right)\left(x-x_{k}\right)=0,
$$

so

$$
x_{k+1}=x_{k}-D f\left(x_{k}\right)^{-1} f\left(x_{k}\right) .
$$

In practice inverses are difficult to calculate (require to many operations) and the linear system for $\Delta x_{k}=x_{k+1}-x_{k}$

$$
D f\left(x_{k}\right) \Delta x_{k}=-f\left(x_{k}\right)
$$

is solved at each step (using $L U$ decomposition of $\operatorname{Df}\left(x_{k}\right)$ ) and hence

$$
x_{k+1}=x_{k}+\Delta x_{k} .
$$

## Example

Derive Newton's method for solving the system of quadratic equations:

$$
\begin{aligned}
& x^{2}+y^{2}-10 x+y=1 \\
& x^{2}-y^{2}-x+10 y=25
\end{aligned}
$$

We are searching for the zero of the vector function

$$
F: \mathbb{R}^{2} \rightarrow \mathbb{R}^{2}, \quad F(x, y)=\left(x^{2}+y^{2}-10 x+y-1, x^{2}-y^{2}-x+10 y-25\right) .
$$

The Jacobian of $F$ in $(x, y)$ is

$$
D F(x, y)=\left[\begin{array}{cc}
2 x-10 & 2 x-1 \\
2 y+1 & -2 y+10
\end{array}\right]
$$

Using Newton's metod we:

- Choose an initial term ( $x_{0}, y_{0}$ ).
- Calculate $x_{r+1}=x_{r}+\Delta x_{r}$, where $D F\left(x_{r}, y_{r}\right) \Delta x_{r}=-F\left(x_{r}, y_{r}\right)^{T}$.

