

COMPUTATIONAL MATHEMATICS

Based on prof. Antonio Frangioni's and
prof. Federico Poloni's lectures

Gemma Martini

August 9, 2020

A sincere thank you to
Alessandro Cudazzo, Donato Meoli,
Giulia Volpi and all those who helped me
improving these notes in style and contents.

Contents

1	20th of September 2018 — F. Poloni	5
1.1	A brief journey in Linear Algebra	5
1.1.1	Matrix multiplications: four flavors plus one of looking at it	10
1.2	A warm up	11
1.3	Solving Linear Systems	14
1.4	Orthogonality	17
2	26th of September 2018 — F. Poloni	19
2.1	Orthogonality (II)	19
2.2	Eigenvalues / Eigenvector	20
2.2.1	Eigenvector: what could possibly go wrong?	23
3	28th of September 2018 — F. Poloni	27
3.1	Singular value decomposition (SVD)	27
3.1.1	Properties of SVD	29
4	4th of October 2018 — F. Poloni	33
5	10th of October 2018 — F. Poloni	35
6	18th of October 2018 — F. Poloni	37
6.1	Least squares problem	37
6.1.1	Method of normal equations	38
6.2	QR factorization	39
7	26th of October 2018 — F. Poloni	43
7.1	How to construct a QR factorization	43
7.1.1	Matlab implementation	46
7.2	How to use the thin QR factorization to solve a least squares problem	47
8	7th of November 2018 — F. Poloni	49
8.1	Least squares problem with SVD	50
8.1.1	Behaviour in case of zeros as singular values	53
8.2	Truncated SVD	54

8.3	Tikhonov regularization / ridge regression	54
9	9th of November 2018 — F. Poloni	57
9.1	Conditioning	57
9.1.1	Conditioning of linear systems	61
9.2	Condition number, SVD, and distance to singularity	62
9.3	Conditioning of least squares problem	63
10	15th of November 2018 — F. Poloni	67
10.1	Stability of algorithms	67
10.2	Backward stability of QR factorization	71
10.2.1	Stability of algorithms for least-squares problems	72
10.2.2	A posteriori checks	74
10.3	A posteriori check for Least Squares Problems	75
11	21st of November 2018 — F. Poloni	77
11.1	Gaussian elimination and LU factorization	78
11.1.1	Stability of LU	81
11.1.2	Gaussian elimination on sparse matrices	82
12	23rd of November 2018 — F. Poloni	85
12.1	Gaussian elimination on symmetric matrices	85
12.2	Cholesky factorization	88
12.3	Krylov subspace methods	89
13	29th of November 2018 — F. Poloni	91
13.0.1	Naive idea	91
13.0.2	Improvement	91
13.1	Arnoldi algorithm	92
14	5th of December 2018 — F. Poloni	97
14.1	Convergence of Arnoldi	98
14.1.1	Better explanation	98
15	7th of December 2018 — F. Poloni	101
16	13th of December 2018 — F. Poloni	105
16.1	Lanczos algorithm	105
17	19th of September 2018	109
17.1	Introduction to machine learning problems	109
17.2	Optimization	111
17.2.1	Linear estimation	111
17.2.2	Low-rank approximation	112
17.2.3	Support vector machines	113

18 21st of September 2018	115
18.1 Mathematical background for optimization problems	115
18.1.1 Multi-objective Optimization	116
18.2 Infima, suprema and extended reals	118
18.3 (Monotone) Sequences in \mathbb{R} and optimization	119
18.4 Vector spaces and topology	120
18.5 Limit of a sequence in \mathbb{R}^n	123
19 27th of September 2018	125
19.1 Continuity	128
19.2 Derivatives	128
19.2.1 Multivariate differentiability	129
20 3rd of October 2018	133
20.1 Simple functions	137
20.1.1 Linear functions	137
20.1.2 Quadratic functions	137
21 5th of October 2018	139
21.1 Unconstrained optimization	139
21.1.1 First order model	140
21.1.2 Second order model	141
21.2 Convexity	142
21.2.1 Convex sets	142
21.2.2 Convex functions	145
22 11th of October 2018	149
22.0.1 Non-differentiable functions	152
23 17th of October 2018	155
23.1 Optimization algorithms	155
23.1.1 Gradient method for quadratic functions	156
24 19th of October 2018	161
24.0.1 Gradient method for quadratic functions (cont.)	161
24.0.2 Quadratic gradient descent in Matlab	164
25 24th of October 2018	167
25.1 Gradient method for non quadratic functions	167
25.1.1 Finding the best step size	168
26 25th of October 2018	173
27 8th of November 2018	185
27.1 Good practices for the project's design	185

28 14th of November 2018	187
28.1 General descent methods	187
28.2 Newton's method	188
28.2.1 Convex case	188
28.2.2 Interpretation of Newton's method	191
28.2.3 Non convex case	192
29 16th of November 2018	195
29.1 Quasi-newton methods	196
29.1.1 Davidson-Fletcher-Powell	196
29.1.2 Broyden-Fletcher-Goldfarb-Shanno	197
29.1.3 Poorman's approach	197
29.2 Conjugate gradient method	198
30 22nd of November 2018	201
30.1 Deflected gradient methods	201
30.1.1 Heavy ball gradient method	201
30.1.2 Accelerated gradient	202
30.2 Incremental gradient methods	204
30.3 Subgradient methods	205
31 28th of November 2018	207
31.1 Subgradient methods	207
31.1.1 Target level stepsize	209
31.2 Deflected subgradient	210
31.3 Smoothed gradient methods	211
31.4 Cutting-plane algorithm	213
31.5 Bundle methods	215
32 30th of November 2018	217
32.1 Constrained optimization	217
32.1.1 Linear equality constraints	218
32.1.2 Background for linear inequality constraints	220
33 6th of December 2018	227
33.1 Duality	227
33.2 Lagrangian duality	228
33.3 Specialized dual	230
33.3.1 Linead programs	230
33.3.2 Quadratic programs	230
33.3.3 Conic program	231
33.4 Fenchel's duality	232

34 12th of December 2018	233
34.1 Quadratic problem with linear equality constraints	233
34.2 Inequality constrained problems	235
34.2.1 Projected gradient method	235
34.2.2 Projected gradient method with box constraints	238
34.2.3 Active-set method for quadratic programs	239
34.2.4 Frank-Wolfe method	240
35 14th of December 2018	241
35.1 Dual methods for linear constrained optimization	241
35.1.1 Separable problems and partial dual	242
35.2 Primal/dual methods or barrier methods	242
35.2.1 Barrier function and central path	242
35.3 Primal-dual interior point method	245

Lecture 1

20th of September 2018 — F. Poloni

In the following chapters we will going to cover the part of Numerical Analysis of the course of Computational Mathematics for Data Learning and Analytics. This part of the course is held by Professor Federico Poloni. The content of these chapters is based primarily on the material provided by the professor on the e-learning portal. In addition to that material, the content has been written also based on the content of the course Linear Algebra held by Prof. Gilbert Strang from MIT. There is a nice portal called MIT OpenCourseWare where you can find a lot of recorded lectures from MIT. Linear algebra course is just super duper fantastic.

During the course, some Matlab code will be covered. We do not plan to include a deep introduction to the Matlab language chiefly for two reasons:

1. there is already plenty of stuff online,
2. it is too time consuming to create one.

1.1 A brief journey in Linear Algebra

What is linear algebra? Linear algebra solves systems of linear equations. Period. But there is a whole world there in. So better start right now.

What is a system of equations? Well, simply put, it is a bunch of linear equations, say m , each of which is characterized by at most n unknowns, i.e. variables. Since there are two dimensions involved here, that is m and n , there are actually two points of view to look at this system. We can look at it from the row perspective and from the column perspective. Let us go through an example to show this.

Example 1.1.1. Suppose we are given the following system of two linear equa-

tions with two unknowns:

$$\begin{cases} -4x + 2y = 2 \\ 3x + 5y = -3 \end{cases}$$

As we probably know, we can write this system of linear equations in the following matrix form:

$$\begin{bmatrix} -4 & 2 \\ 3 & 5 \end{bmatrix} \begin{bmatrix} x \\ y \end{bmatrix} = \begin{bmatrix} 2 \\ -3 \end{bmatrix}$$

Usually the coefficient matrix is indicated with A , the vector of unknowns is indicated with x , and the vector of coefficients on the right hand side is usually indicated with b .

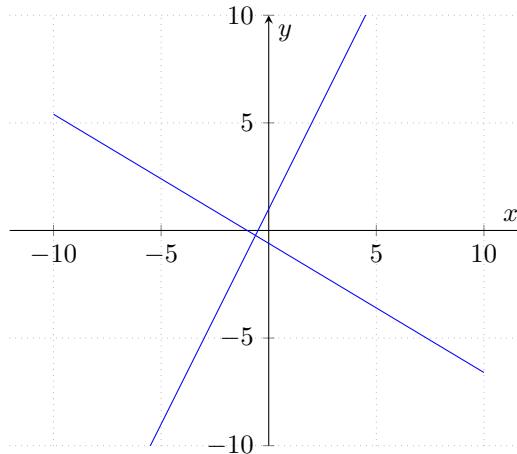
The row perspective is the one that we are mostly used to deal with. Each equation in this example represents a line in the cartesian plane and the problem requires to find the point (if it exists) where the two lines meet (see figure 1.1).

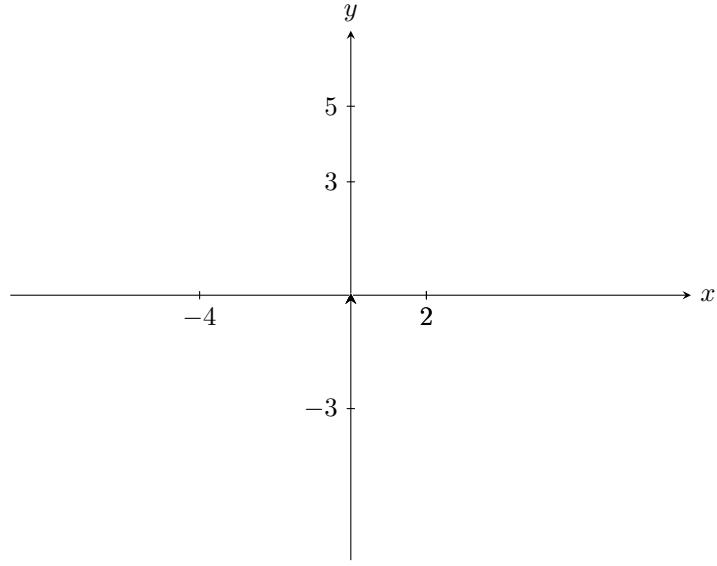
The column perspective is a bit different and maybe completely new to you. Look at this mind blowing thing:

$$x \begin{bmatrix} -4 \\ 3 \end{bmatrix} + y \begin{bmatrix} 2 \\ 5 \end{bmatrix} = \begin{bmatrix} 2 \\ -3 \end{bmatrix}$$

Why is column perspective interesting? Because now we can look at these columns of A as vectors in a column space. We are eager to find the right amount (unknowns) of each of these column vectors that produce the vector b (see figure 1.1).

Linear combination is the fundamental operation of the whole course. Besides the question of what is the linear combination of the columns of A that produce b , we will also ask ourselves what are all the possible vectors b that we can obtain with some linear combination of the columns of A ? Or what are all the possible linear combinations of the columns of A that give a certain b ?





In the previous example we have encountered a bunch of new terms. Let us now define them rigorously. We saw that the generic system of linear equations can be written as:

$$Ax = b$$

Note that we are actually abusing terminology here. Even though it is not that important, bare in mind that these are actually affine equations. The linear equations are the equations that go through the origin, while affine need not to.

We have seen that there is a row perspective and a column perspective. These perspectives are actually two different Euclidean spaces. The first is called row space and the second one is the column space. Note how these correspond directly to the rows and columns of A respectively. When the matrix is a square matrix the dimension of the two spaces is the same (if certain properties hold). But when we have $m \times n$ matrices, then the two spaces have different dimensions (if the same properties hold). Remember all these spaces, because it is important! Especially the column space.

Clearly, two equations in two unknowns is a very basic case. Let us see another basic case before moving to interesting stuff :).

Example 1.1.2. Let us now take three equations in three unknowns. *TODO: I need a tikz magician here...*

Now that we have said that the linear system of equations can also be seen as a bunch of vectors in the n dimensional space, we can think of solving a linear system of equations as the problem of finding a linear combination of columns of the matrix A to obtain the vector b . Thus, given a vector b , how can we obtain it (if it is possible at all) as a linear combination of the columns of A ? If you think a bit, the vector b is yet another vector in the n dimensional space. Think

of the columns of A as directions, i.e. n directions. You are in the origin of the space, and you want to get to the point pointed by b . You may move in each of n directions only once and in every direction you can choose how much you want to move. You may move backwards (corresponding to negative coefficient in the linear combination) or forward (corresponding to positive coefficient in the linear combination). A zero coefficient means that you do not move in the corresponding direction. The problem is: given these n direction, can you get from the origin to the point b ? Note, whatever b you are given. This basically means that these n direction must somehow “fill” the whole space, that is for every point b there must be a linear combination of directions that bring you from the origin to b . So, is there a way to know whether a given set of columns is “sufficiently powerful” to be able to generate all the “treasury maps”? Here comes the concept of linear independence.

If the columns of A are linearly independent, then every vector b on the right hand side of the equation $Ax = b$ can be generated with some combination of columns of A . What does it mean to be linearly independent? It means that when you take one of the n columns, you cannot find a linear combination of the remaining $n - 1$ columns of A that can generate the one that you have picked. Geometrically speaking, $n - 1$ columns in an n dimensional space generate a so called subspace. In an n dimensional space there are n types of subspaces: a zero subspace, the subspaces of 1 dimension, the subspaces of 2 dimensions, ..., the subspaces of $n - 1$ dimensions, and finally a subspace of n dimensions. Note singularity and plurality in the previous sentence. There is only one subspace of zero dimension, namely the origin. There is only one subspace of n dimensions in an n dimensional space, the space it self. The remaining subspaces are all of dimensions 1 to $n - 1$. A subspace of dimension 1 is a line. A subspace of dimension 2 is a plane. A subspace of dimension 3 is the three dimensional space. A subspace of dimension 4... well you picture it, I'm not able. So $n - 1$ vectors in the n dimensional space generate a subspace (think of a plane subspace in the three dimensional space). If the vector that we have picked from A is located on the very same plane generated by the $n - 1$ vectors, then there is no way that these $n - 1$ vectors can generate a vector that is outside of this plane, because there is no direction that points outside of the plane itself. The vector that we have picked and that is on the plane is then linearly dependant from the $n - 1$ remaining ones. Clearly, we may have more than one vectors that are linearly dependant from the rest in A . The obvious question now is what is the subspace that the columns of A generate? In other words what is the maximum number of linearly independent columns of A ? In linear algebra this is the so called rank of the matrix. When the rank is n then the matrix is said to be full rank. Given this definition, the rank of the matrix then satisfy the following box constraint:

$$0 \leq \text{rank}(A) \leq n$$

for some $A \in M(n, n, \mathbb{R})$. Since we have two dimensions, namely rows and columns, each of which generates its own space, namely row space and column space, you may legitimately wonder if these are the same. Well, in case of square matrices $n \times n$ the two are the same. So the rank of row space is the same of

the column space. You want a proof? Search rank of a matrix on wikipedia for three different proofs :).

Let us now turn back to the rows. The question for you is: what happens if we do instead:

$$x^T A = b^T$$

where x^T is the row vector, so we have:

$$\begin{bmatrix} x_1 & x_2 & \dots & x_n \end{bmatrix} \begin{bmatrix} a_{11} & a_{12} & \dots & a_{1n} \\ a_{21} & a_{22} & \dots & a_{2n} \\ \vdots & \vdots & \dots & \vdots \\ a_{n1} & a_{n2} & \dots & a_{nn} \end{bmatrix} = [b_1 \ b_2 \ \dots \ b_n]$$

The row vector b is the linear combination of the rows of A where the weights are given by the vector x . Namely:

$$x_1 [a_{11} \ a_{12} \ \dots \ a_{1n}] + \dots + x_n [a_{n1} \ a_{n2} \ \dots \ a_{nn}] = [b_1 \ b_2 \ \dots \ b_n]$$

Everything we have said about columns, linear independence, rank, etc, holds for rows and row space of the matrix A . Basically, when we do operations from the left of the matrix A then we are operating on rows of the matrix, when instead we are operating from the right then we are performing operations over columns of A .

Now that we know that linearly independent columns of A can generate all the vectors in the corresponding n dimensional column space, the question is how do we generate it? In other words, given a vector b how do we find the coefficients of the linear combination that generates it? Note the linear combination and not a linear combination. There is only one linear combination that generates it. Why? Picture the row space. Each equation in the linear system generates a subspace of dimension $n - 1$. So in three dimensional space the equations represent planes. In case of a system of n linear equations in n unknowns, we have n subspaces of dimension $n - 1$ in the n dimensional row space. So if these n subspaces are linearly independent, then it must be the case that there is only one point that touches them all when they intersect. If there is more than one point that is in the intersection of all of them, then the matrix A is not full rank. Suppose then that the columns (or rows) are linearly independent, how do we find the coefficients? Note this basic math:

$$ax = b \iff \frac{1}{a}x = \frac{b}{a} \iff x = a^{-1}b$$

where every term is a simple scalar. So to get x we must multiply b with the inverse of a . The same happens with the matrices and vectors. We need something called inverse matrix of A , indicated with A^{-1} :

$$x = A^{-1}b$$

The matrix A is invertible only if it is full column rank. This is why the linear independence is so central to the whole linear algebra. Note that only square

matrices may have the inverse matrix. Obviously, the inverse matrix is unique. And moreover the inverse of the inverse is the initial matrix, namely:

$$(A^{-1})^{-1} = A$$

Cool man. But how the heck are we going to find the inverse? There are many (not *that* many) algorithms that can be used for matrix inversion. But one of the standard one is the Gauss-Jordan Elimination algorithm. Are we going to cover it here? Why not? It is one of the most used algorithms in linear algebra so let us give it some love. But before doing this algorithm, we still need one ingredient: matrix multiplications.

1.1.1 Matrix multiplications: four flavors plus one of looking at it

Matrix multiplication is another of the most important operations that you can think of when you think of linear algebra. In here, we are going to cover briefly four different ways to look at this important operation.

We start with the basic definition that you have probably seen if you have ever done some linear algebra course. The classical definition is:

$$\begin{bmatrix} a_{11} & a_{12} & \dots & a_{1n} \\ a_{21} & a_{22} & \dots & a_{2n} \\ \vdots & \vdots & \dots & \vdots \\ a_{n1} & a_{n2} & \dots & a_{nn} \end{bmatrix} \begin{bmatrix} b_{11} & b_{12} & \dots & b_{1n} \\ b_{21} & b_{22} & \dots & b_{2n} \\ \vdots & \vdots & \dots & \vdots \\ b_{n1} & b_{n2} & \dots & b_{nn} \end{bmatrix} = \begin{bmatrix} c_{11} & c_{12} & \dots & c_{1n} \\ c_{21} & c_{22} & \dots & c_{2n} \\ \vdots & \vdots & \dots & \vdots \\ c_{n1} & c_{n2} & \dots & c_{nn} \end{bmatrix}$$

where:

$$c_{ij} = \sum_{k=1}^n a_{ik} b_{kj}$$

But as we all know, there is more than one of everything. Another way to look at matrix matrix multiplication is the following. We have seen that matrix vector multiplication is basically a linear operation that send a certain vector x into another vector y . Thus, $Ax = y$ means that x is mapped to y through A . When it comes to the matrix matrix multiplication, i.e. $AB = C$, the same story happens. It is the columns of B that are mapped to columns of C . Each column i of B is mapped to the corresponding column i of C . Similarly, we could think of all of this from the row space perspective. The rows of A are mapped to the corresponding rows of C through B . So these are the additional two ways to look at matrix matrix multiplication. What about the forth one?

The forth way to look at matrix matrix multiplication is a bit more mind blowing. In the first way, we have computed the dot products between rows of A and columns of B . But nobody prevents us from doing the contrary. What if

we used the columns of A and the rows of B :

$$\begin{bmatrix} a_{1i} \\ a_{2i} \\ \vdots \\ a_{ni} \end{bmatrix} \begin{bmatrix} b_{i1} & b_{i2} & \dots & b_{in} \end{bmatrix}$$

What do we get for each $i = [1 \dots n]$? We get an $n \times n$ matrix. So the resulting matrix C is the summation of n different matrices of size $n \times n$.

1.2 A warm up

Before we start, here is a small recap of the basic linear algebra operations that we will use throughout the whole course:

- **Vector-Scalar product:**

Let $x \in \mathbb{R}^n$ and $\lambda \in \mathbb{R}$ we call **multiple** of vector x the following vector:

$$\lambda x = x\lambda = \begin{pmatrix} \lambda x_1 \\ \vdots \\ \lambda x_n \end{pmatrix}$$

- **Vector-Vector product:**

Let $x, y \in \mathbb{R}^n$, the dot product between these two vectors is defined as $x^T y = \sum_{i=1}^n x_i y_i$. Note in particular that $x^T y \in \mathbb{R}$. In other words, the dot product produces a scalar.

- **Scalar-Matrix product:**

Let $A \in M(n, m, \mathbb{R})$ and $\lambda \in \mathbb{R}$ we call the **scalar-matrix product** the following:

$$\lambda A = A\lambda = \begin{pmatrix} \lambda A_{11} & \lambda A_{12} & \dots & \lambda A_{1m} \\ \lambda A_{21} & \lambda A_{22} & \dots & \lambda A_{2m} \\ \vdots & \vdots & \ddots & \\ \lambda A_{n1} & \lambda A_{n2} & \dots & \lambda A_{nm} \end{pmatrix}$$

- **Matrix-Vector product:**

Given a matrix $A \in M(n, m, \mathbb{R})$ and a vector $v \in \mathbb{R}^m$ the **matrix-vector product** $Av = w \in \mathbb{R}^n$ is computed as follows:

$$w = Av = \begin{pmatrix} A_1 v \\ A_2 v \\ \vdots \\ A_m v \end{pmatrix}, w_i = \sum_{j=1}^m A_{ij} v_j$$

This is the simple way, just a row-by-column vector product, the computational complexity of this operation is $O(n^2)$.

The smart way to compute it: **linear combinations** of columns of A, e.g.:

$$\begin{pmatrix} A_{11} & A_{12} & A_{13} \\ A_{21} & A_{22} & A_{23} \\ A_{31} & A_{32} & A_{33} \\ A_{41} & A_{42} & A_{43} \end{pmatrix} \begin{pmatrix} v_1 \\ v_2 \\ v_3 \end{pmatrix} = \begin{pmatrix} w_1 \\ w_2 \\ w_3 \end{pmatrix}$$

with **linear combinations** we have:

$$\begin{pmatrix} A_{11} \\ A_{21} \\ A_{31} \\ A_{41} \end{pmatrix} v_1 + \begin{pmatrix} A_{12} \\ A_{22} \\ A_{32} \\ A_{42} \end{pmatrix} v_2 + \begin{pmatrix} A_{13} \\ A_{23} \\ A_{33} \\ A_{43} \end{pmatrix} v_3 = \begin{pmatrix} w_1 \\ w_2 \\ w_3 \\ w_4 \end{pmatrix}$$

- **Matrix-Matrix Product:**

Given two matrices $A \in M(n, m, R)$ and $B \in M(m, k, R)$ we call **matrix-matrix product** the following: $C = AB$ such that $C_{ij} = A_i B^j$, where $A_i^T \in \mathbb{R}^m$ is the i -th row of A , B^i is the i -th column of B ($B^i \in \mathbb{R}^m$) and $C \in M(n, k, \mathbb{R})$. Notice that this product is **not commutative**: $AB \neq BA$ might not even make sense dimension-wise.

As long as the complexity is concerned, multiplying two matrices $m \times n$ and $n \times k$ requires $O(mnk)$ floating point operations (flops). Forget about fancier algorithms (e.g. Strassen)

Order of operations

Usual algebra properties hold, e.g.: $A(B+C) = AB+AC$, $A(BC) = (AB)C, \dots$

Parenthesization matters a lot: if $A, B \in M(n, \mathbb{R})$, $v \in \mathbb{R}^n$, then $(AB)v$ costs $O(n^3)$, but $A(Bv)$ costs $O(n^2)$. Programming languages usually do not rearrange parentheses to help.

- **Image** of a matrix A ($\text{Im}(A)$): the set of vectors that can be obtained multiplying A by any vector in the domain of A .
- **Kernel** of a matrix A ($\text{ker}(A)$): the set of vectors w in its domain such that $Aw = 0$.
- Given a matrix $A \in M(n, \mathbb{R})$ we call **inverse** of A the matrix A^{-1} such that:

$$A^{-1}A = AA^{-1} = I_n = \begin{pmatrix} 1 & & & & \\ & 1 & & & \\ & & 1 & & \\ & & & \ddots & \\ & & & & 1 \end{pmatrix}$$

The **inverse of a product** (shoe-sock identity) is $(AB)^{-1} = B^{-1}A^{-1}$. Notice that this identity holds only for square matrices.

- The **transpose** of a matrix $A \in M(n, m, \mathbb{R})$ is A^T such that $A_{ij}^T = A_{ji}$. The **transpose of a product** (shoe-sock identity) is $(AB)^T = B^T A^T$. (This identity holds for square and rectangular matrices)

Definition 1.2.1. *General linear group (GL): the general linear group of degree n is the set of $n \times n$ invertible matrices, together with the operation of ordinary matrix multiplication*

Fact 1.2.1. *Let $A \in GL(n, \mathbb{R})$ (aka A is a real square matrix of size n and invertible), $B, C \in M(n, m, \mathbb{R})$ and we have the equality $AB = AC$. If there is a matrix M such that $MA = I$, the following holds*

$$(MA)B = (MA)C \iff B = C, M = A^{-1}$$

In general, $AB = AC$ does not imply $B = C$; it holds only when A is invertible.

Row and column vectors notation

$$v = \begin{pmatrix} 4 \\ 5 \\ 6 \end{pmatrix}, v^T = (4 \quad 5 \quad 6)$$

v is a column vector in \mathbb{R}^3 (or a matrix in $M(3, 1, \mathbb{R})$) and v^T is a row vector (or a matrix in $M(1, 3, \mathbb{R})$).

Definition 1.2.2 (Basis). *We call **basis** a set B of elements (vectors) in a vector space V if every element of V may be written in a unique way as a (finite) linear combination of elements of B . The coefficients of this linear combination are referred to as components or coordinates on B of the vector. The elements of a basis are called **basis vectors**.*

Definition 1.2.3 (Canonical basis). *We term **canonical basis** of a vector space \mathbb{R}^n the basis made of all the column of the $n \times n$ identity matrix $I_n \in M(n, \mathbb{R})$.*

Example 1.2.1. In \mathbb{R}^4 the **canonical basis** is $\mathcal{B} = \{e_1, e_2, e_3, e_4\}$ such that

$$e_1 = \begin{pmatrix} 1 \\ 0 \\ 0 \\ 0 \end{pmatrix}, e_2 = \begin{pmatrix} 0 \\ 1 \\ 0 \\ 0 \end{pmatrix}, e_3 = \begin{pmatrix} 0 \\ 0 \\ 1 \\ 0 \end{pmatrix}, e_4 = \begin{pmatrix} 0 \\ 0 \\ 0 \\ 1 \end{pmatrix}$$

and each vector $w \in \mathbb{R}^4$ can be written as $w = w_1e_1 + w_2e_2 + w_3e_3 + w_4e_4$.

The powerful idea behind linear algebra: many relations are true regardless of the basis we use. E.g. w , v and $w + v$ in two different bases.

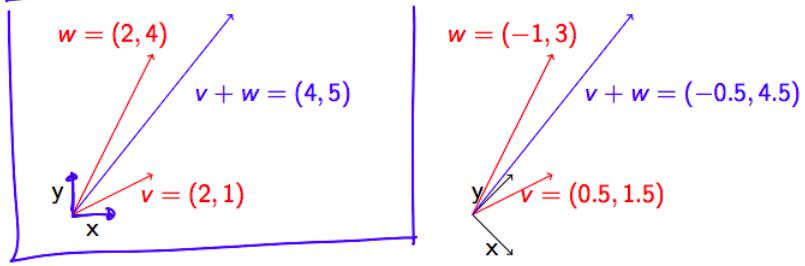


FIGURE 1.1: How a change of basis reflects on the space.

Definition 1.2.4 (Triangular matrix). Let $A \in M(n, \mathbb{R})$. We term A **upper triangular** if $(A)_{ij} = 0$ for each $i < j$. Conversely, we term A **lower triangular** if $(A)_{ij} = 0$ for each $j < i$. The set of all triangular $n \times n$ real matrices is a group and it is denoted as $T(n, \mathbb{R})$.

Definition 1.2.5 (Diagonal matrix). Let $A \in M(n, \mathbb{R})$. We term A **diagonal** if $(A)_{ij} = 0$ for each $i \neq j$. The set of all diagonal $n \times n$ real matrices is a group and it is denoted as $D(n, \mathbb{R})$.

Definition 1.2.6 (Symmetric matrix). Let $A \in M(n, \mathbb{R})$. We term A **symmetric** if $(A)_{ij} = (A)_{ji}$ for each $i, j = 1, \dots, n$. The set of all diagonal $n \times n$ real matrices is a group and it is denoted as $D(n, \mathbb{R})$.

1.3 Solving Linear Systems

The objective of this course, for the part concerning numerical methods, is solving linear systems efficiently.

Definition 1.3.1 (Linear system). Let $A \in M(n, m, \mathbb{R})$, $b \in \mathbb{R}^n$ and $x \in \mathbb{R}^m$. We term **linear system** the following:

$$Ax = b$$

Our goal is to *approximate* such vector x , hence resulting in solving a minimum problem:

$$\min \|Ax - b\|$$

If we have a square and invertible matrix $A \in GL(n, \mathbb{R})$ solving a linear system means: find those coordinates x_1, \dots, x_n needed to write b as a linear combination of the columns of (square) A and in this case, the solution is given by: $x = A^{-1}b$.

Warning: this is not the best way to solve a linear system on a computer!



Something on Matlab ...

Notice that the machine precision is 10^{-16} , so we should pay attention when making computations, since we may incur in some error (proportional to the size of the operands).

In Matlab a matrix is written as $A=[1, 2, 3; 4, 5, 6];$, where $[1, 2, 3]$ is the first row of the matrix A.

The transpose of a matrix or a vector is denoted by A' .

The inverse of a square matrix is denoted by $\text{inv}(A)$.

If we are interested in only a part of our matrix A we may write $A(1:2, 1:3)$ and obtain only the rows of A that go from 1 to 2 and those columns from 1 to 3.

Notice that in Matlab both vector and matrices are 1-based.

Definition 1.3.2 (Block multiplications). *Let $A \in M(n, m, \mathbb{R})$ and let $B \in M(m, k, \mathbb{R})$. We can compute the result of a block of the matrix AB as the product of the two blocks in A and B in the corresponding position.*

Observation 1.3.1. *When computing a matrix product, we get the same result if we use the row-by-column rule block-wise.*

$$\left(\begin{array}{ccc|cc} \times & \times & \times & \times & \times & \times \\ \times & \times & \times & \times & \times & \times \\ \hline \times & \times & \times & \times & \times & \times \\ \times & \times & \times & \times & \times & \times \\ \hline \times & \times & \times & \times & \times & \times \\ \times & \times & \times & \times & \times & \times \end{array} \right) \cdot \left(\begin{array}{ccc|cc} \times & \times & \times & \times & \times & \times \\ \times & \times & \times & \times & \times & \times \\ \hline \times & \times & \times & \times & \times & \times \\ \times & \times & \times & \times & \times & \times \\ \hline \times & \times & \times & \times & \times & \times \\ \times & \times & \times & \times & \times & \times \end{array} \right) = \left(\begin{array}{ccc|cc} \times & \times & \times & \times & \times & \times \\ \times & \times & \times & \times & \times & \times \\ \hline \times & \times & \times & \times & \times & \times \\ \times & \times & \times & \times & \times & \times \\ \hline \times & \times & \times & \times & \times & \times \\ \times & \times & \times & \times & \times & \times \end{array} \right)$$

$$\left(\begin{array}{ccc} \times & \times & \times \\ \times & \times & \times \\ \times & \times & \times \end{array} \right) \cdot \left(\begin{array}{ccc} \times & \times & \times \\ \times & \times & \times \\ \times & \times & \times \end{array} \right) + \left(\begin{array}{cc} \times & \times \\ \times & \times \end{array} \right) \cdot (\times \quad \times \quad \times) + \left(\begin{array}{cc} \times & \times \\ \times & \times \end{array} \right) \cdot \left(\begin{array}{ccc} \times & \times & \times \\ \times & \times & \times \\ \times & \times & \times \end{array} \right) = \left(\begin{array}{ccc} \times & \times & \times \\ \times & \times & \times \end{array} \right)$$

Notice that block operations usually give better performance: one matrix-matrix product performs faster than n matrix-vector products (even if they have the same number of flops). This is one of the reasons why library calls usually perform better than hand-coded loops (Blas/Lapack).

Fact 1.3.1 (Block triangular matrices). *Let $M \in M(n, m, \mathbb{R})$ and $B \in M(m, k, \mathbb{R})$ such that they are **block triangular**. Their product is a block triangular matrix as well. In other words, block triangular matrices are closed under products:*

$$MB = \begin{pmatrix} A & B \\ 0 & C \end{pmatrix} \begin{pmatrix} D & E \\ 0 & F \end{pmatrix} = \begin{pmatrix} AD & AE + BF \\ 0 & CF \end{pmatrix}$$

Fact 1.3.2 (Properties of block triangular matrices).

Let M be a block triangular matrix, where all the blocks on the diagonal are square

$$M = \begin{pmatrix} A_{11} & A_{12} & \cdots & A_{1n} \\ 0 & A_{22} & \cdots & A_{2n} \\ \vdots & \ddots & \ddots & \vdots \\ 0 & \cdots & 0 & A_{nn} \end{pmatrix}$$

1. A block triangular matrix is invertible iff all diagonal blocks A_{ii} are invertible;
2. The eigenvalues of a block triangular matrix are the union of the eigenvalues of each A_{ii} block;
3. Let $M \in GL(n, \mathbb{R})$ such that $M = \begin{pmatrix} A & B \\ 0 & C \end{pmatrix}$ the inverse of M is

$$M^{-1} = \begin{pmatrix} A^{-1} & -A^{-1}BD^{-1} \\ 0 & C^{-1} \end{pmatrix}.$$
4. The product of two block (upper/lower) triangular matrices (with compatible block sizes) is still block triangular

Why are we interested in block triangular matrices? They depict a situation as shown in Figure 1.2.

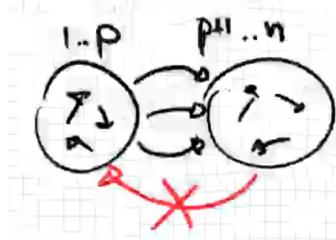


FIGURE 1.2: The adjacency matrix of a bipartite graph has 0s in its bottom left part (Matlab syntax $A[p+1:n; 1:p]=0$), which means that the edges from a connected component and the other are in one direction only.

General principle: matrix structures matter. Block triangular linear system has a cheaper system solution than a general system as shown in Example 1.3.1.

Example 1.3.1. Let us take a 2×2 block triangular linear system

$$\begin{pmatrix} A & B \\ 0 & C \end{pmatrix} \cdot \begin{pmatrix} x \\ y \end{pmatrix} = \begin{pmatrix} e \\ f \end{pmatrix}$$

(Again, diagonal blocks are square and all dimensions are compatible.)

$$\begin{pmatrix} Ax + By \\ Cy \end{pmatrix} = \begin{pmatrix} e \\ f \end{pmatrix} \implies y = C^{-1}f, x = A^{-1}(e - BC^{-1}f)$$

$$\begin{pmatrix} A & B \\ 0 & C \end{pmatrix}^{-1} = \begin{pmatrix} A^{-1} & -A^{-1}BC^{-1} \\ 0 & C^{-1} \end{pmatrix}$$

Informal idea: we can start solving from the variables multiplied by C.

1.4 Orthogonality

Definition 1.4.1 (Norms). Let $x \in \mathbb{R}^n$. We “measure” its magnitude using so-called “norms”.

EUCLIDEAN: $\|x\|_2 = x^T x = \sqrt{\sum_{i=1}^n x_i^2}$;

NORM 1: $\|x\|_1 = \sum_{i=1}^n |x_i|$;

p -NORM: $|x|_p = \left(\sum_{i=1}^n |x_i|^p \right)^{1/p}$;

0-NORM: $\|x\|_0 = |\{i : |x_i| > 0\}|$, which accounts for 1 – #of 0 entries;

∞ -NORM: $\|x\|_\infty = \max_{i=1,\dots,n} |x_i|$.

From now on in this part of the course, if not explicitly specified, we will refer to norm-2 only.

Definition 1.4.2 (Scalar product). Let $v, w \in \mathbb{R}^n$ we term **standard scalar product** between v and w the real number $v^T w = \sum_{i=1}^n v_i w_i$.

Definition 1.4.3 (Orthogonal vectors). Let $v, w \in \mathbb{R}^n$. We say that v is **orthogonal** to w (in symbols $x \perp y$) if their scalar product is zero.

Formally, $x^T v = 0$.

Definition 1.4.4 (Orthogonal matrix). Let $A \in M(n, \mathbb{R})$ a square matrix. We term A **orthogonal** if $A^T A = A A^T = I_n$ where I_n is the identity matrix of size n (1 on the diagonal, 0 elsewhere) or equivalently if $A^{-1} = U^T$.

The set of all orthogonal matrices in $M(n, \mathbb{R})$ is a group and it called orthogonal group and denotes as $O(n, \mathbb{R})$.

Fact 1.4.1. Let $A \in O(n, \mathbb{R})$, $\forall x \in \mathbb{R}^n$ we have that $\|Ax\| = \|x\|$.

Proof.

$$\|Ax\| = \sqrt{(Ax)^T Ax} = \sqrt{x^T A^T Ax} = \sqrt{x^T I_n x} = \sqrt{x^T x} = \|x\|$$

□

Lecture 2

26th of September 2018 —
F. Poloni

2.1 Orthogonality (II)

In the previous lecture we introduced some sufficient conditions for matrix orthogonality.

Theorem 2.1.1. Let $U \in O(n, \mathbb{R})$ and let $x \in \mathbb{R}^n$. Then $\|Ux\| = \|x\|$.

Proof. Instead of proving that $\|Ux\| = \|x\|$ we will prove $\|Ux\|^2 = \|x\|^2$:

$$\|Ux\|^2 = (Ux)^T \cdot (Ux) \stackrel{(1)}{=} x^T U^T U x = x^T I_n x = x^T x = \|x\|^2$$

where $\stackrel{(1)}{=}$ follows from the definition of transpose of a product. \square

Geometrically an orthogonal transformations (aka matrices) preserve the norm, so an orthogonal matrix A represents a symmetry or a rotation these operations do not alter the size of vectors.

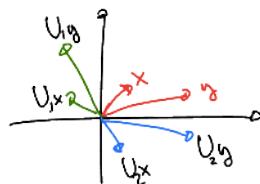


FIGURE 2.1: Matrix U_1 represents a rotation, while U_2 is a symmetry operation.

Definition 2.1.1 (Orthonormality). Let $x, y \in \mathbb{R}^n$ we say that x and y are **orthonormal** if $\langle x, y \rangle = 0$ and $\|x\| = \|y\| = 1$.

Fact 2.1.2. Let us take $U \in \mathbb{O}(n, \mathbb{R})$. Then its columns U^1, U^2, \dots, U^n are **orthonormal** and the same holds for its rows.

$$U^{i^T} U^j = \begin{cases} 1 & \text{if } i = j \\ 0 & \text{otherwise} \end{cases}$$

and

$$U_i U_j^T = \begin{cases} 1 & \text{if } i = j \\ 0 & \text{otherwise} \end{cases}$$

Fact 2.1.3. The set of orthogonal matrices is closed under product operations. Let $U, V \in \mathbb{O}(n, \mathbb{R})$, then $U \cdot V$ is orthogonal.

$$\text{Proof. } (UV)^T \cdot (UV) = V^T U^T U V = V^T I_n V = V^T V = I_n \quad \square$$

Since we will often deal with tall-thin rectangular matrices with orthonormal columns as U_1

$$U_1 = (U^1 \quad U^2 \quad \dots \quad U^n) \in M(m, n, \mathbb{R}) \text{ where } m \geq n$$

the following fact may come in handy

Fact 2.1.4. $\forall U_1 \in M(m, n, \mathbb{R})$ where $M \geq n$ and the columns of U_1 are orthogonal $\exists U_2$ s.t. $(U_1 \quad U_2) \in \mathbb{O}(m, \mathbb{R})$.

2.2 Eigenvalues / Eigenvector

Definition 2.2.1 (Eigenvectors and eigenvalues). Let $A \in M(n, \mathbb{R})$ and let $x \neq 0 \in \mathbb{R}^n$ and $\lambda \in \mathbb{R}$.

If $Ax = \lambda x$ we say that x is an **eigenvector** of **eigenvalue** λ .

Theorem 2.2.1. Let $A \in M(n, \mathbb{R})$ (real triangular matrix). The eigenvalues of A are the scalars on the diagonal.

Terminology

In this notes we refer to the columns of a generic matrix A as A^1 for the first column, A^2 for the second column and so on and so forth.

Conversely, we represent the rows of a matrix A as A_i .

Definition 2.2.2 (Diagonalizable matrix). Some matrices $A \in M(n, \mathbb{R})$ under some conditions can be decomposed as:

$$A = V \Lambda V^{-1}$$

$$A = V \Lambda V^{-1} = \begin{pmatrix} V^1 & V^2 & \dots & V^n \end{pmatrix} \cdot \begin{pmatrix} \lambda_1 & & 0 & \\ & \lambda_2 & & \\ 0 & & \ddots & \\ & & & \lambda_n \end{pmatrix} \cdot \begin{pmatrix} w_1 \\ w_2 \\ \vdots \\ w_n \end{pmatrix}$$

where $V \in GL(n, \mathbb{R})$ is the matrix that has as columns the eigenvectors of A (V^i, λ_i) $\forall i = 1, \dots, n$ and $w_i = V_i^{-1}$ are the rows of the inverse of matrix V .

Fact 2.2.2. $\forall V^i \in \mathbb{R}$ eigenvectors of A they are still eigenvector of the diagonalized form of $A = V\Lambda V^{-1}$

Proof.

$$AV^i = V\Lambda V^{-1}V^i = V\Lambda e_i = \lambda_i V^i$$

□

Another way to see the diagonalized form of A is the following:

$$A = V\Lambda V^{-1} = \sum_{i=1}^n v_i \lambda_i w_i^T =$$

$$\boxed{v_1} \cdot \boxed{\lambda_1} \cdot \boxed{w_1^T} + \boxed{v_2} \cdot \boxed{\lambda_2} \cdot \boxed{w_2^T} + \cdots + \boxed{v_n} \cdot \boxed{\lambda_n} \cdot \boxed{w_n^T}$$



Something on Matlab ...

Notice that in Matlab the eigenvalues and eigenvectors of a matrix are computed using the command `[V, Lambda] = eig(U)` and this operation has a computational complexity of $O(n^3)$.

We can check that the matrix A is equal to the decomposition in this way:
`A - V*Lambda*inv(V)` or `norm(A - V*Lambda*inv(V))` (both should be close to zero).

Notice that not all matrices $A \in M(n, \mathbb{R})$ allow a diagonal decomposition. It may happen that such a matrix A is diagonalizable in \mathbb{C} and its eigenvalues are complex.

This decomposition tell us the behavior under repeated application of a matrix A to a vector x . This process allow to scale a general vector x .



Something on Matlab ...

e.g `A = [1 1; 1 1]` and `x = [1 1]`
then `A*x` is equal to `[2 2]'` and `A*A*x` is equal to `[4 4]'`

Fact 2.2.3. If $A \in M(n, \mathbb{R})$ is diagonalizable (aka may be written as $A = V\Lambda V^{-1}$) then $A^k x = \sum_{i=1}^n \lambda_i^k \alpha_i V^i$, for some $\alpha_i \in \mathbb{R}$.

Proof. Let us write x in the base of \mathbb{R}^n made of the linearly independent columns of V :

$$x = V^1\alpha_1 + V^2\alpha_2 + \cdots + V^n\alpha_n$$

for some $\alpha_i \in \mathbb{R}$.

- ALGEBRAIC VIEW POINT:

$$\begin{aligned} Ax &= A \cdot (V^1\alpha_1 + V^2\alpha_2 + \cdots + V^n\alpha_n) \\ &= AV^1\alpha_1 + AV^2\alpha_2 + \cdots + AV^n\alpha_n \\ &= \lambda_1 V^1\alpha_1 + \lambda_2 V^2\alpha_2 + \cdots + \lambda_n V^n\alpha_n \\ &= V^1(\lambda_1\alpha_1) + V^2(\lambda_2\alpha_2) + \cdots + V^n(\lambda_n\alpha_n) \end{aligned} \tag{2.2.1}$$

Then

$$\begin{aligned} A^2x &= A \cdot \left(V^1(\lambda_1\alpha_1) + V^2(\lambda_2\alpha_2) + \cdots + V^n(\lambda_n\alpha_n) \right) \\ &= AV^1\lambda_1\alpha_1 + AV^2\lambda_2\alpha_2 + \cdots + AV^n\lambda_n\alpha_n \\ &= \lambda_1^2 V^1\alpha_1 + \lambda_2^2 V^2\alpha_2 + \cdots + A\lambda_n^2 V^n\alpha_n \end{aligned} \tag{2.2.2}$$

The thesis follows inductively.

- LINEAR ALGEBRA VIEW POINT:

$$\begin{aligned} A^kx &= A \cdot A \cdot \cdots \cdot A \cdot x \\ &= V\Lambda V^{-1}V\Lambda V^{-1}\cdots V\Lambda V^{-1}x \\ &= V\Lambda^k V^{-1}x \\ &= V \begin{pmatrix} \lambda_1^k & & & \\ & \lambda_2^k & & \\ & & \ddots & \\ & & & \lambda_n^k \end{pmatrix} V^{-1}x \\ &= V \begin{pmatrix} \lambda_1^k & & & \\ & \lambda_2^k & & \\ & & \ddots & \\ & & & \lambda_n^k \end{pmatrix} \cdot \begin{pmatrix} \alpha_1 \\ \alpha_2 \\ \vdots \\ \alpha_n \end{pmatrix} \end{aligned} \tag{2.2.3}$$

□

Notice that if A is not square, $Av, \lambda v$ have different sizes and it doesn't make sense to talk about eigenvalues.

2.2.1 Eigenvector: what could possibly go wrong?

1. The eigenvalue decomposition is highly non-unique, we can:
 - Reorder eigenvalues/vectors
 - Replace an eigenvector v_i with $2v_i$, $3.5v_i \dots$
 - For matrices with repeated eigenvalues we have even more possibilities:
e.g $I = VIV^{-1}$ for every invertible V
2. some matrices have only complex eigenvalues: e.g. $\begin{pmatrix} 2 & 4 \\ -3 & 3 \end{pmatrix}$
3. some matrices have fewer eigenvectors than we want and we can't use eigenvalue decomposition: e.g. $\begin{pmatrix} 1 & 1 \\ 0 & 1 \end{pmatrix}$

Now, thanks to the eigenvalue decomposition we can prove the following:

Theorem 2.2.4. Let $A \in M(n, \mathbb{R})$. If $|\lambda_i| < 1$ for all eigenvalues λ_i of A then $\lim_{k \rightarrow \infty} A^k x = 0$.

Theorem 2.2.5. Let $A \in M(n, \mathbb{R})$. If $\forall \lambda_i$ eigenvalues of A $|\lambda_i| < |\lambda_1|$ then $A^k x \approx V^1 \lambda_1^k \alpha_1$.

Fact 2.2.6. Let $A \in M(n, \mathbb{R})$ be a diagonalizable matrix and let

$$A = V\Lambda V^{-1} = \begin{pmatrix} V^1 & V^2 & \dots & V^n \end{pmatrix} \cdot \begin{pmatrix} \lambda_1 & & & \\ & \lambda_2 & & \\ & & \ddots & \\ & & & \lambda_n \end{pmatrix} \cdot \begin{pmatrix} w_1 \\ w_2 \\ \vdots \\ w_n \end{pmatrix}$$

Let us now consider a reordering of V 's columns and apply the same permutation to the “diagonal vector” of Λ such that

$$\hat{V} = \begin{pmatrix} V^2 & V^1 & V^3 \dots & V^n \end{pmatrix} \text{ and } \hat{\Lambda} = \begin{pmatrix} \lambda_2 & & & \\ & \lambda_1 & & \\ & & \lambda_3 & \\ & & & \ddots \\ & & & & \lambda_n \end{pmatrix}$$

A can be diagonalized through such \hat{V} and $\hat{\Lambda}$ as $A = V\Lambda V^{-1} = \hat{V}\hat{\Lambda}\hat{V}^{-1}$.

Moreover, in the case of repeated eigenvalues

Fact 2.2.7. Let $A \in M(n, \mathbb{R})$ a diagonalizable matrix such that $A = V\Lambda V^{-1}$, where $\lambda_1 = \lambda_2$ (without loss of generality). Then V can be replaced by $\tilde{V} = \begin{pmatrix} & & & & \\ V^1 + V^2 & V^1 - V^2 & V^3 & \dots & V^n \end{pmatrix}$.

Theorem 2.2.8 (Spectral theorem). *Let $A \in S(n, \mathbb{R})$ (A is a real symmetric matrix). Then A is diagonalizable $A = U\Lambda U^{-1}$, where eigenvalues are all real numbers and we can take U orthogonal matrix.*

Notice that for symmetric matrices none of the “unlucky” situations enumerated above may happen (it is justified by the spectral theorem).



Something on Matlab ...

If we have an symmetric matrix B and we compute $[V, D] = \text{eig}(B)$, matlab will always return an orthogonal matrix V .

Definition 2.2.3 (Quadratic form). *Let $Q \in S(n, \mathbb{R})$ we define **quadratic form***

$$f(x) = x^T Q x$$

for each $x \in \mathbb{R}^n$.

Geometrically, a quadratic form defines a paraboloids.

Example 2.2.1. Let $f_1(\begin{pmatrix} x_1 \\ x_2 \end{pmatrix}) = (x_1 \ x_2) \begin{pmatrix} 3 & 2 \\ 2 & 4 \end{pmatrix} \begin{pmatrix} x_1 \\ x_2 \end{pmatrix} = 3x_1^2 + 4x_1x_2 + 4x_2^2$
and let $f_2(\begin{pmatrix} x_1 \\ x_2 \end{pmatrix}) = (x_1 \ x_2) \begin{pmatrix} 3 & 2 \\ 2 & -4 \end{pmatrix} \begin{pmatrix} x_1 \\ x_2 \end{pmatrix} = 3x_1^2 + 4x_1x_2 - 4x_2^2$
For a graphic hint see Figure 2.2.

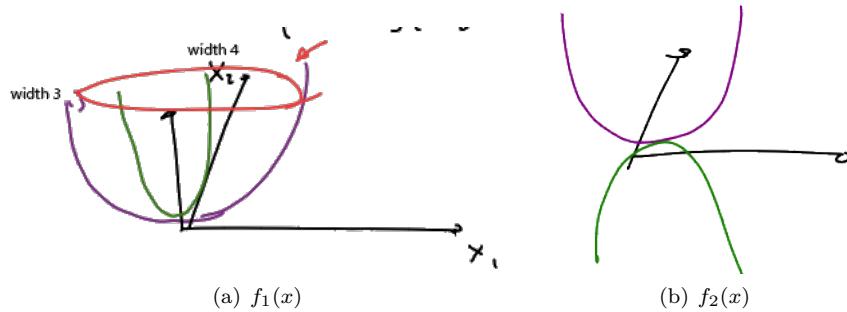


FIGURE 2.2: The plot of functions.

Fact 2.2.9. *Let $Q \in S(n, \mathbb{R})$ (For a fixed symmetric matrix) and let $x \in \mathbb{R}^n$. Then*

$$\lambda_{\min}\|x\|^2 \leq x^T Q x \leq \lambda_{\max}\|x\|^2$$

where λ_{\max} and λ_{\min} are respectively the eigenvalue of maximum value and the eigenvalue of minimum value.

Proof. EASY CASE WITH Q DIAGONAL:

$$x^T Q x = x^T \begin{pmatrix} \lambda_2 & & & \\ & \lambda_1 & & \\ & & \lambda_3 & \\ & & & \ddots \\ & & & & \lambda_n \end{pmatrix} x = \lambda_1 x_1^2 + \lambda_2 x_2^2 + \cdots + \lambda_n x_n^2$$

It is obvious that this sum is bounded by:

$$\lambda_{\min} \cdot (x_1^2 + x_2^2 + \cdots + x_n^2) \leq \lambda_1 x_1^2 + \lambda_2 x_2^2 + \cdots + \lambda_n x_n^2 \leq \lambda_{\max} \cdot (x_1^2 + x_2^2 + \cdots + x_n^2)$$

The following holds: $\lambda_{\min} \cdot (x_1^2 + x_2^2 + \cdots + x_n^2) = \lambda_{\min} \cdot x^T x = \lambda_{\min} \cdot \|x\|^2$ and, on the other hand, $\lambda_{\max} \cdot (x_1^2 + x_2^2 + \cdots + x_n^2) = \lambda_{\max} \cdot x^T x = \lambda_{\max} \cdot \|x\|^2$ and this proves the fact in the special case of diagonal matrix Q .

GENERAL CASE: Let us represent Q through its eigendecomposition: $A = U \Lambda U^{-1} = U \Lambda U^T$, where U is an orthogonal matrix.

$$x^T Q x = x^T U \Lambda U^T x \stackrel{(1)}{=} y^T \Lambda y$$

where $\stackrel{(1)}{=}$ is due to the change of variable $y = U^T x$ (that implies $y^T = x^T U$).

By the same argument used in the diagonal case,

$$\lambda_{\min} \cdot \|y\|^2 \leq y^T \Lambda y \leq \lambda_{\max} \cdot \|y\|^2$$

Now the point is that if we can replace $\|U^T x\|^2 = \|y\|^2$ with $\|x\|^2$ we have proved the theorem. In fact this is true, due to the orthogonality of matrix U and Proposition 1.4.1.

□

Corollary 2.2.10. *Let $Q \in S(n, \mathbb{R})$ and let $x \in \mathbb{R}^n$. If $x \neq 0$, $\lambda_{\min} \leq \frac{x^T Q x}{\|x\|^2} \leq \lambda_{\max}$, where λ_{\max} and λ_{\min} are respectively the eigenvalue of maximum value and the eigenvalue of minimum value.*

Definition 2.2.4 (Positive semidefinite). *Let $Q \in S(n, \mathbb{R})$. We say that Q is positive semidefinite (and we indicate $\succeq 0$) if*

$$x^T Q x \geq 0 \quad \|x\|^2 \geq 0$$

Definition 2.2.5 (Positive definite). *Let $Q \in S(n, \mathbb{R})$. We say that Q is positive definite (and we indicate $\succ 0$) if*

$$x^T Q x \geq 0 \quad \|x\|^2 > 0$$

Fact 2.2.11. Let $Q \in S(n, \mathbb{R})$. The following holds

$$Q \text{ is positive semidefinite} \iff \forall \lambda \text{ eigenvalue of } Q \quad \lambda \geq 0$$

And this holds with the > 0 in the positive definite case.

Proof.

(\Leftarrow) $x^T Q x \geq \lambda_{\min} \geq 0$ since we are in the hypothesis that all the eigenvalues are ≥ 0
 $\Rightarrow \forall v_i \text{ eigenvector of } Q \quad 0 \leq v_i^T A v_i = v_i^T \cdot (\lambda_i v_i) = \lambda_i v_i^T v_i = \lambda_i \|v_i\|^2 \Rightarrow \lambda_i \geq 0$. \square

Corollary 2.2.12. Let $Q \in S(n, \mathbb{R})$ such that $Q \succeq 0$. The following holds:

$$A \text{ is invertible} \iff Q \text{ is strictly positive definite}$$

Fact 2.2.13. Let $B \in M(m, n, \mathbb{R})$ (possibly rectangular), $B^T B \in S(n, \mathbb{R})$ is a valid product and it is a square, symmetric, positive semidefinite matrix.

Proof. SYMMETRY: $(B^T B)^T = B^T \cdot (B^T)^T = B^T B$.

POSITIVE DEFINITE: $x^T B^T B x = (Bx)^T (Bx) = \|Bx\|^2 \geq 0$

\square

Corollary 2.2.14. The same holds for $BB^T \in S(m, \mathbb{R})$ and it is easily proved taking $C = B^T \in M(n, m, \mathbb{R})$.



Something on Matlab ...

In order to check if a matrix A is positive definite in Matlab we can look at its eigenvalues (cfr. `eig(A)`).

Notice that in the **complex** case most of these properties work as well, but it is needed to replace A^T with $\overline{A^T}^*$ (transpose and entrywise conjugate). The norm of a complex vector is computed as

$$\|x\|_2^2 = x^* x = \overline{x_1} x_1 + \overline{x_2} x_2 + \cdots + \overline{x_n} x_n = |x_1|^2 + \cdots + |x_n|^2 \in \mathbb{R}^+ \cup \{0\}$$

Moreover, in the complex case, $UU^* = I$ is called **unitary matrix** (orthogonal + complex)

Often denoted with A^ or A^H and called **Hermitian matrix**.

Lecture 3

28th of September 2018 — F. Poloni

3.1 Singular value decomposition (SVD)

We are left with the task of reaching a (sort of) “eigenvalue decomposition” when the target matrix is not symmetric.

There are two ways to generalize the eigenvalue decomposition to a nonsymmetric matrix A (with something that always exists):

Definition 3.1.1 (Schur decomposition). *Let $A \in M(n, \mathbb{R})$, $\exists U \in M(n, \mathbb{R})$ orthogonal matrix and $T \in M(n, \mathbb{R})$ triangular matrix such that $A = UTU^T$ and this is called **Schur decomposition**.*

What is really important for us is the **Singular value decomposition**, every square matrix A can be written with **SVD** form.

Every square matrix A can be written as SVD.

Definition 3.1.2 (Singular value decomposition). *Let $A \in M(n, \mathbb{R})$, $\exists U, V \in M(n, \mathbb{R})$ orthogonal matrices (V not necessarily equal to U) and $\Sigma \in Diag(n, \mathbb{R})$ such that $A = U\Sigma V^T$ and this is called **Singular Value Decomposition**.*

$$A = (u_1 \ u_2 \ \cdots \ u_n) \cdot \begin{pmatrix} \sigma_1 & & & \\ & \sigma_2 & & \\ & & \ddots & \\ & & & \sigma_n \end{pmatrix} \cdot \begin{pmatrix} {v_1}^T \\ {v_2}^T \\ \vdots \\ {v_n}^T \end{pmatrix} = \sum_{i=1}^n u_i \sigma_i {v_i}^T = \\ = u_1 \sigma_1 {v_1}^T + u_2 \sigma_2 {v_2}^T + \cdots + u_n \sigma_n {v_n}^T$$

Where σ_i are called **singular values** and they are sorted such that:

$$\sigma_1 \geq \sigma_2 \geq \cdots \geq \sigma_n \geq 0$$

General fact on singular values:

- *Singular values \neq eigenvalues*
- *They are always positive and usually more spread apart than the eigenvalues.*

$$\sigma_1 \geq |\lambda_1| \text{ and } |\lambda_m| \geq \sigma_m$$

λ_i is larger than the largest eigenvalue of a matrix A and λ_m is smaller than the smallest eigenvalue of a matrix A .

The SVD can be defined also for a rectangular matrix A :

Definition 3.1.3 (Rectangular matrices and SVD). *Let $A \in M(m, n, \mathbb{R})$, there exist $U \in M(m, \mathbb{R})$ orthogonal, $V \in M(n, \mathbb{R})$ orthogonal and $\Sigma(m, n, \mathbb{R})$ diagonal in the sense that $\sum_{ij} = 0$ with $i \neq j$ (padded with zeros). Matrix A has a **SVD factorization**, where Σ has the following shape:*

- case $m < n$ (e.g $m = 3, n = 5$)

$$\begin{pmatrix} \sigma_1 & 0 & 0 & 0 & 0 \\ 0 & \sigma_2 & 0 & 0 & 0 \\ 0 & 0 & \sigma_3 & 0 & 0 \end{pmatrix}$$

- case $m > n$ (e.g $m = 5, n = 3$)

$$\begin{pmatrix} \sigma_1 & 0 & 0 \\ 0 & \sigma_2 & 0 \\ 0 & 0 & \sigma_3 \\ 0 & 0 & 0 \\ 0 & 0 & 0 \end{pmatrix}$$

Definition 3.1.4 (Thin SVD). *Let $A \in M(m, n, \mathbb{R})$, has a **thin SVD factorization**: we may restrict to compute only the first $\min(m, n)$ vectors that appear in this sum: thin SVD.*

$$A = \sum_{i=1}^{\min(m,n)} u_i \sigma_i v_i^T = u_1 \sigma_1 v_1^T + u_2 \sigma_2 v_2^T + \cdots + u_{\min(m,n)} \sigma_{\min(m,n)} v_{\min(m,n)}^T$$

Something on Matlab ...

In Matlab the SVD decomposition is obtained through the command `svd(A)`, which return value is made of the three matrices U, Σ, V .

As an example, `[U, S, V] = svd(A)`. Notice that, if `svd(A)` is assigned to one variable, then such variable is an array of singular values.

The thin SVD can be compute with: `[U, S, V] = svd(A, 0)`

Computational costs

We are not going into details of algorithms for computing SVD, but we would like to add a consideration about the computational complexity of such an algorithm.

- $[U, S, V] = \text{svd}(A, 0)$ (thin) costs $O(mn^2)$ ops for $A \in \mathbb{R}^{m \times n}$ or $A \in \mathbb{R}^{n \times m}$ with $m \geq n$
- $[U, S, V] = \text{svd}(A)$ (non-thin) more expensive, because it has to store the large $m \times m$ factor. (But there are some tricks to store orthogonal matrices compactly, more about it later).

3.1.1 Properties of SVD

The SVD reveals rank, image, and kernel of a matrix.

Definition 3.1.5 (Rank). *Let $A \in M(n, \mathbb{R})$ we call the **rank** of A the number of non-zero singular values.*

Equivalently, the rank is the size of the column space.

Property 3.1.1. *A matrix $A \in M(n, \mathbb{R})$ has rank r iff all its eigenvalues starting from the $r+1$ -th are 0, formally iff $\sigma_1 \geq \sigma_2 \geq \dots \geq \sigma_r > \sigma_{r+1} = \dots = \sigma_n = 0$.*

Thanks to Property 3.1.1, we can somehow talk about an “even thinner” SVD, where all the 0s in the bottom right part of the matrix Σ , cancel out the latter columns of U and the latter rows of V (aka columns of V^T). A pictorial representation of the shape of Σ can be found below.

$$\Sigma = \begin{pmatrix} \sigma_1 & & & \\ & \sigma_2 & & \\ & & \ddots & \\ & & & \sigma_r \\ \hline & & & 0 \\ & & & & \ddots \\ & & & & & 0 \\ \hline & & & & & \mathbf{0} \end{pmatrix}$$

This factorization represents A as $\sum_{i=1}^r U_i \sigma_i V_i$.

An attentive reader may notice that $Ax = \sum_{i=1}^r U_i \sigma_i V_i x$, where the last three terms are dimensionally a scalar. It goes without saying that the image of A is the span of U_1, U_2, \dots, U_r , hence $\text{rk}(A) = r$.

Moreover, $\ker(A) = \text{span}(V_{r+1}, V_{r+2}, \dots, V_n)$, since V is orthogonal (proof: plugging in $x = V_j$, where $j > r$).

Definition 3.1.6 (Matrix norm). *Let $A \in M(m, n, \mathbb{R})$. We define the **matrix norm** of A as*

$$\|A\| := \max_{x \neq 0} \frac{\|Ax\|}{\|x\|} = \max_{z=1} \|Az\|$$

Where the norm may be any of the ones defined in Definition 1.4.1 second equality is introduced in order to work in a compact set, the one of normalized vectors z .

Notice that $\|Ax\| \leq \|A\| \cdot \|x\|$.

Property 3.1.2. Let A and $B \in M(n, m, \mathbb{R})$ and let $x \in \mathbb{R}^n$, the following holds, for any norm defined in Definition 1.4.1:

- $\|A\| \geq 0$ (and the equality holds iff $A = 0$);
- $\|\alpha A\| = |\alpha| \|A\|, \forall \alpha \in \mathbb{R}$;
- $\|A + B\| \leq \|A\| + \|B\|$;
- $\|AB\| \leq \|A\| \|B\|$;
- $\|Ax\| \leq \|A\| \|x\|$.

Fact 3.1.3. Let $A \in (n, m, \mathbb{R})$ and let $U \in M(m, n, \mathbb{R})$ orthogonal, in the case of 2-norm $\|A\|_2 = \|AU\|_2 = \|UA\|_2$.

Proof. $\|UA\|_2 = \max_{x \in R^n, x \neq 0} \frac{\|UAx\|_2}{\|x\|_2} \stackrel{(1)}{=} \max_{x \in R^n, x \neq 0} \frac{\|Ax\|_2}{\|x\|_2} = \|A\|_2$, where $\stackrel{(1)}{=}$ follows from a property of vector norms.

$\|AU\|_2 = \max_{x \in R^n, x \neq 0} \frac{\|AUx\|_2}{\|x\|_2} \stackrel{(2)}{=} \max_{y \in R^m, y \neq 0} \frac{\|Ay\|_2}{\|y\|_2} = \|A\|_2$, where $\stackrel{(2)}{=}$ follows from the substitution $y = Ux$. \square

Definition 3.1.7 (Frobenius norm). Let $A \in M(n, m, \mathbb{R})$, we term **Frobenius norm** of A $\|A\|_F = \sqrt{\sum_{i=1}^n \sum_{j=1}^m (A)_{ij}^2}$.

Notice that all the properties enumerated in Property 3.1.2 hold for the Frobenius norm as well.

Fact 3.1.4. Let $A \in M(n, m, \mathbb{R})$ and let $A = U\Sigma V^T$ be its singular value decomposition. The following hold:

1. $\|A\|_2 = \|\Sigma\|_2 = \sigma_1$

2. $\|A\|_F = \|Sigma\|_F = \sum_{i=1}^{\min n, m} \sigma_i^2$

Proof. 1. The first equality follows from Proposition 3.1.3, while the second

is proved as follows:

$$\begin{aligned}
\|\Sigma\|_2 &= \max_{x \in \mathbb{R}^n, x \neq 0} \frac{\|\Sigma x\|_2}{\|x\|_2} = \max_{x \in \mathbb{R}^n, x \neq 0} \frac{\left\| \begin{pmatrix} \sigma_1 & & & \\ & \sigma_2 & & \\ & & \ddots & \\ & 0 & & \sigma_n \end{pmatrix} \cdot \begin{pmatrix} x_1 \\ x_2 \\ \vdots \\ x_n \end{pmatrix} \right\|_2}{\left\| \begin{pmatrix} x_1 \\ x_2 \\ \vdots \\ x_n \end{pmatrix} \right\|_2} = \max_{x \in \mathbb{R}^n, x \neq 0} \frac{\left\| \begin{pmatrix} \sigma_1 x_1 \\ \sigma_2 x_2 \\ \vdots \\ \sigma_n x_n \\ 0 \\ \vdots \\ 0 \end{pmatrix} \right\|_2}{\left\| \begin{pmatrix} x_1 \\ x_2 \\ \vdots \\ x_n \end{pmatrix} \right\|_2} \\
&= \frac{\sqrt{(\sigma_1 x_1)^2 + (\sigma_2 x_2)^2 + \cdots + (\sigma_n x_n)^2}}{\sqrt{x_1^2 + x_2^2 + \cdots + x_n^2}} \leq \frac{\sqrt{(\sigma_1 x_1)^2 + (\sigma_1 x_2)^2 + \cdots + (\sigma_1 x_n)^2}}{\sqrt{x_1^2 + x_2^2 + \cdots + x_n^2}} \\
&= \sqrt{\sigma_1^2} \cdot \frac{\sqrt{x_1^2 + x_2^2 + \cdots + x_n^2}}{\sqrt{x_1^2 + x_2^2 + \cdots + x_n^2}} = \sigma_1
\end{aligned} \tag{3.1.1}$$

The equality is achieved if we pick $x = e_1 = \begin{pmatrix} 1 \\ 0 \\ \vdots \\ 0 \end{pmatrix}$.

2. The proof of this assertion is similar to the other and it is left to the reader. \square

Theorem 3.1.5 (Eckart-Younger). *Let $A \in M(n, m, \mathbb{R})$ and let $A = U\Sigma V^T$ be its singular value decomposition.*

The solution of $\min_{rk(X) \leq k} \|A - X\|$ is given by the truncated SVD:

$$X = (U^1 \ U^2 \ \cdots \ U^k) \cdot \begin{pmatrix} \sigma_1 & & & \\ & \sigma_2 & & \\ & & \ddots & \\ & & & \sigma_k \end{pmatrix} \cdot \begin{pmatrix} V^1 \\ V^2 \\ \vdots \\ V^k \end{pmatrix}$$

Where the norm is both $\|\cdot\|_2$ and $\|\cdot\|_F$.

Fact 3.1.6. *Let $A \in M(n, \mathbb{R})$ and let A be invertible. The following holds: $\|A^{-1}\| = \frac{1}{\sigma_n}$*

Proof. Since A is invertible, none of the σ_i is 0, hence the smaller (namely σ_n) is not 0.

$$A^{-1} = (U\Sigma V^T)^{-1} \stackrel{(1)}{=} V^T \Sigma^{-1} U^{-1} = V \begin{pmatrix} \frac{1}{\sigma_1} & & & \\ & \frac{1}{\sigma_2} & & \\ & & \ddots & \\ & & & \frac{1}{\sigma_n} \end{pmatrix} U^T$$

Where $\stackrel{(1)}{=}$ follows from the orthogonality of V and U .

Notice that this is *almost* an SVD, because the values on the diagonal are not sorted in a decreasing order.

Plugging in the norm, we have:

$$\|A^{-1}\| = \left\| V \begin{pmatrix} \frac{1}{\sigma_1} & & & \\ & \frac{1}{\sigma_2} & & \\ & & \ddots & \\ & & & \frac{1}{\sigma_n} \end{pmatrix} U^T \right\| = \left\| \begin{pmatrix} \frac{1}{\sigma_1} & & & \\ & \frac{1}{\sigma_2} & & \\ & & \ddots & \\ & & & \frac{1}{\sigma_n} \end{pmatrix} \right\| = \frac{1}{\sigma_n}$$

□

Lecture 4

4th of October 2018 — F. Poloni

This lecture is about practical usage of the singular value decomposition and takes place almost wholly on Matlab.

For example, given a certain image, that can be represented as a matrix of values in the range [0, 255], the rank-1 SVD of such image, results in a very abstract picture, see Figure 4.1. The more we increase the rank, the better is the similarity of the approximated image with respect to the original one.



Something on Matlab ...

Given a certain matrix A , we can compute the SVD decomposition using the command `[U, S, V] = svd(A)`

Definition 4.0.1 (Principal component analysis). *Given a matrix A , we term **principal component analysis** the analysis of features of such matrix via the rows and columns of U and V respectively, where U and V are the matrices of the SVD decomposition.*

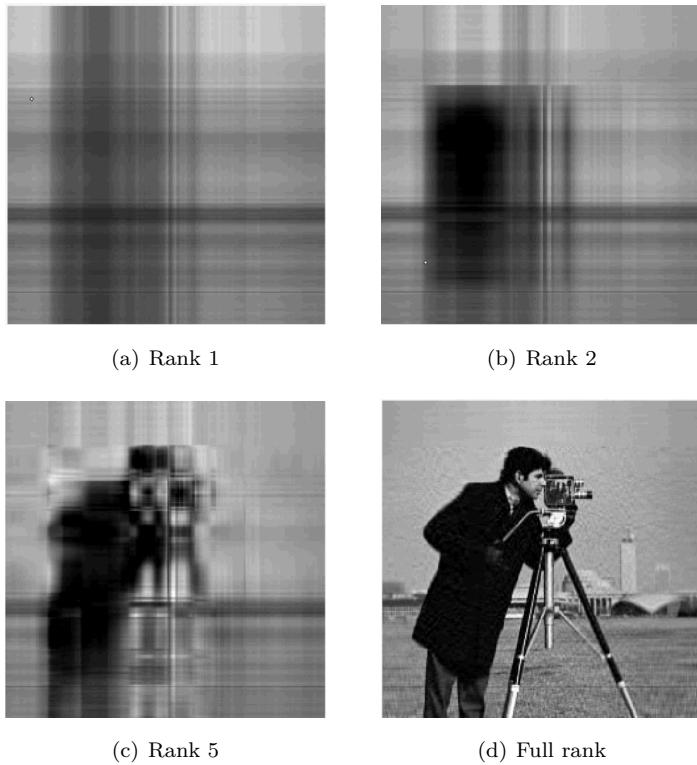


FIGURE 4.1: How the approximation of a matrix changes with respect to the different ranks.

Lecture 5

10th of October 2018 — F. Poloni

This lecture has the goal of introducing the concept of linear combinations.

Definition 5.0.1 (Linear combination). *In a very unformal way, we can define the goal of linear combination as the pursuit of obtaining a certain target vector $b \in \mathbb{R}^n$ using m (in principle $m \neq n$) vectors a_1, a_2, \dots, a_m such that*

$$a_1x_1 + a_2x_2 + \cdots + a_mx_m = b$$

where x_i are properly chosen.

The task of finding such vectors is called **solving a linear system** and it is formally written as $Ax = b$.

Theorem 5.0.1. *Let $A \in M(n, m)$ and let $b \in \mathbb{R}^n$. It holds that any linear system $Ax = b$ is solvable iff A is invertible.*

We are interested in finding approximate solutions of such systems, where the proximity to the target is expressed in terms as $\|Ax - b\|$ that should be close to zero. A geometric intuition is displayed in Figure 5.1.



Something on Matlab ...

Matlab provides syntactic sugar to solve linear systems.

Before introducing such syntax let us just notice the following $5 \setminus 2$
 $(= 2/5) \neq 5/2$.

The syntax to solve $Ax = b$ is $A \setminus b$, where the algorithm used in Matlab is not inverting the matrix A and then performing the multiplication, but it is a more sophisticated and efficient one.

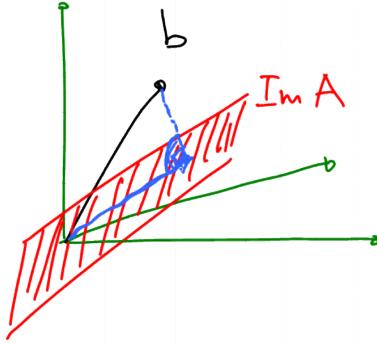


FIGURE 5.1: In this case the image of the matrix A (in red) does not contain b and the best one can do is to obtain a projection of b in the plane $\text{Im}(A)$ (drawn in blue).

Definition 5.0.2 (Linearly square problem). *Let $A \in M(n, m, \mathbb{R})$ and let $b \in \mathbb{R}^n$, we term **linearly square problem** the task of computing $\min_{x \in \mathbb{R}^m} \|Ax - b\|_2$.*



Something on Matlab ...

In Matlab the syntax `.func` means that function func should be performed entry by entry of the non-scalar variable.

An example of a practical least square problem may be predicting the salary of NBA players, assuming that the income is obtained as a linear combination of some features.

Definition 5.0.3 (Full rank matrix). *Let $A \in M(n, m, \mathbb{R})$ we say that A has full column rank if $\ker A = \{0\}$.*

Equivalently, $\text{rk}(A) = n$ or alternatively $\nexists z \in \mathbb{R}^n \setminus \{0\}$ such that $Az = 0$.

Fact 5.0.2. *Let $A \in M(n, m, \mathbb{R})$, the least square problem $\|Ax - b\|$ has a unique solution iff A has full column rank.*

Theorem 5.0.3. *Let $A \in M(n, m, \mathbb{R})$. A has full column rank iff $A^T A$ is positive definite.*

Proof. A has full column rank $\iff \|Az\| \neq 0, \forall z \in \mathbb{R}^m \setminus \{0\} \iff \|Az\|^2 \neq 0, \forall z \in \mathbb{R}^m \setminus \{0\} \iff 0 = (Az)^T Az = z^T A^T Az$ \square

Lecture 6

18th of October 2018 — F.
Poloni

6.1 Least squares problem

Fact 6.1.1. Given $A \in \mathcal{M}(m, n, \mathbb{R})$ the following conditions are equivalent:

- $A^T A$ is positive definite;
- A has full column rank;
- the columns of A are linear independent;
- $\text{Ker}(A) = \{0\}$.

Fact 6.1.2. Given $A \in \mathcal{M}(m, n, \mathbb{R})$, if $A^T A$ is positive semidefinite $f(x) = x^T Qx - q^T x + b^T b$ is **strongly** (or **strictly**) convex. In other words, $f(x)$ has a **unique** minimum.

We find the minimum by solving $\nabla f(x) = 0$,

where $f(x) = x^T A^T Ax - 2b^T Ax + b^T b$

$\nabla f(x) = 2A^T Ax - 2A^T b$ How should we compute this gradient?

We know that $f(x + h) = f(x) + (\nabla f(x))^T h + o(\|h\|)$

$$\begin{aligned} f(x + h) &= (x + h)^T A^T A(x + h) - 2b^T A(x + h) + b^T b \\ &= \mathbf{x^T A^T A x} + \mathbf{x^T A^T A h} + \mathbf{h^T A^T A x} + \mathbf{h^T A^T A h} - \mathbf{2b^T A x} - \mathbf{2b^T A h} + \mathbf{b^T b} \\ &= \mathbf{f(x)} + (\mathbf{2x^T A^T A h} - \mathbf{2b^T A h}) + \mathbf{o}(\|\mathbf{h}\|) \\ &= f(x) + (\mathbf{A^T A x} - \mathbf{2A^T b})^T \mathbf{h} + o(\|h\|) \end{aligned} \tag{6.1.1}$$

So, $\nabla \mathbf{f(x)} = \mathbf{A^T A x} - \mathbf{2A^T b}$

We would like to know when the gradient is 0.

$$\nabla f(x) \stackrel{?}{=} 0 \Leftrightarrow A^T A x = A^T b$$

Since $A^T A$ is a **square** matrix and also **non singular** (which means invertible) we may find x by solving a linear system $x = (A^T A)^{-1} (A^T b)$ via:

- Gaussian elimination;
- LU factorization;
- QR factorization;
- Cholesky factorization (specialized method for positive definite matrices)
Idea: $A^T A$ can be written as $A^T A = R^T R$, where R is a square, upper triangular matrix.

Why do we need factorization method? Let's compute complexity:

- $A^T A \rightarrow 2mn^2$, where $m > n$;
- $A^T b \rightarrow 2mn$;
- Solving $A^T A x = A^T b$ with gaussian elimination has a computational complexity of $\frac{2}{3}n^3$;
- Cholesky factorization $A^T A = R^T R$ which has a cost of $\frac{1}{3}n^3$

6.1.1 Method of normal equations

This method solves least squares problem and takes his name from the fact that “normal” means orthogonal.

The key idea is using symmetry to skip half of the entries of $A^T A$

If $Ax = b$ can't be solved, since A is tall and thin, we can multiply on both sides by A^T and try again, since the matrix is square now.

The residual $Ax - b$ is orthogonal to any vector $v \in \text{span}(A)$: $(Av)^T (Ax - b) = 0$

Why?

$$v^T (A^T A x - A^T b) = 0, \text{ see Figure 6.1.}$$

It's possible to find a close formula for solving this problem: $\min \|Ax - b\|$ is given by $x = (A^T A)^{-1} A^T b$.

Definition 6.1.1 (Moore-Penrose pseudoinverse). *Let A be a matrix in $\mathcal{M}(n, m, \mathbb{R})$, the **Moore-Penrose pseudoinverse** of A with full column rank is $\mathbf{A}^\dagger := (A^T A)^{-1} A^T$*

So we can write $x = \mathbf{A}^\dagger b$ for the solution of a LS problem.

In particular, the solution of $\|\min \|Ax - (b_1 + b_2)\|$ is the sum of two solutions of $\min \|Ax_1 - b_1\|$ and $\min \|Ax_2 - b_2\|$.

Obs: $AA^\dagger = A(A^T A)^{-1} A^T = AA^{-1} A^{T-1} A^T = I_{n \times m}$, but this doesn't hold for $A^\dagger A = (A^T A)^{-1} (A^T A)$, which is not the identity matrix.

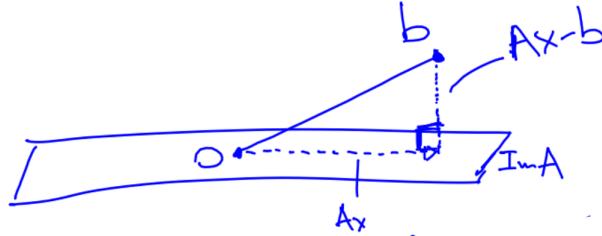


FIGURE 6.1: Geometric idea.

6.2 QR factorization

Theorem 6.2.1. $\forall A \in \mathcal{M}(m, m, \mathbb{R})$, $\exists Q \in \mathcal{O}(m, m, \mathbb{R})$ (space of orthogonal matrices of size $m \times m$), $\exists R \in \mathcal{M}(m, m, \mathbb{R})$ upper triangular such that $A = QR$

QR factorization isn't as powerful as SVD factorization.

Why are we interested in studying factorizations?

- They reveal properties: singularity, rank, ...;
- They may be an intermediate step in algorithms.

Example 6.2.1. We would like to solve $Ax = b$, with $A \in \mathcal{M}(m, m, \mathbb{R})$ we may:

1. first compute the QR factorization ($A = QR$) and then obtain $x = A^{-1}b = R^{-1}Q^{-1}b$
2. compute then $c = Q^T b$
3. and then $x = R^{-1}c$

What's the computational cost?

1. $QR \rightarrow O(m^3)$
2. compute $c \rightarrow O(m^2)$
3. compute $x \rightarrow O(m^2)$

Let's analyze the case in which A is a vector. Given $x \in \mathbb{R}^m$, we want to

find an orthogonal matrix Q such that Qx has the form $\begin{pmatrix} s \\ 0 \\ \vdots \\ 0 \end{pmatrix} = se_1$, where $s = \pm \|x\|$.
 We denote e_i the i -th column of the identity matrix.

Definition 6.2.1 (Householder reflector). Let U be a vector in \mathbb{R}^m . An **Householder reflector** is a matrix H such that $H = I - \frac{2}{U^T U} \cdot UU^T$.

We can observe that $U^T U$ is a scalar.

Since $U^T U = \|U\|^2$, another way of seeing H may be $H = I - \frac{2}{\|U\|^2} UU^T = I - 2vv^T$, where $v = \frac{1}{\|U\|} U$

Lemma 6.2.2. Matrices of this form are orthogonal.

Proof.

$$\begin{aligned}
HH^T &= (I - \frac{2}{\|U\|^2} \cdot UU^T) \cdot (I - \frac{2}{\|U\|^2} \cdot UU^T) \\
&= I \cdot I - \frac{2}{\|U\|^2} \cdot UU^T I - I \cdot \frac{2}{\|U\|^2} \cdot UU^T + \frac{2}{\|U\|^2} \cdot UU^T \cdot \frac{2}{\|U\|^2} \cdot UU^T \\
&= I - \frac{2}{\|U\|^2} \cdot UU^T - \frac{2}{\|U\|^2} \cdot UU^T + \frac{4}{\|U\|^4} \cdot UU^T UU^T \\
&= I - \frac{4}{\|U\|^2} UU^T + \frac{4}{\|U\|^4} U\|U\|^2 U^T \\
&= I
\end{aligned} \tag{6.2.1}$$

□

Example 6.2.2. $Hx = (I - \frac{2}{\|U\|^2} UU^T)x = x - \frac{2}{\|U\|^2} U(U^T U)$

$y = \text{compute_product}(U, x)$

$a = U' * x$

$b = U' * U$

$y = x \frac{2*a}{b} \cdot U$

All these operations are linear operations, so the complexity is $O(m)$, cheaper than generic matrix-vector product ($O(m^2)$).

Can we find a matrix H in this family that maps x to y (equivalently $Hx = y$)?
The answer is given by the following lemma

Lemma 6.2.3. $\forall x, y \text{ s.t. } \|x\| = \|y\| \exists H \text{ s.t. } Hx = y$, choosing $u = x - y$.

A geometric idea is given by the following

What happens if $y = \begin{pmatrix} \|x\| \\ 0 \\ \vdots \\ 0 \end{pmatrix}$?

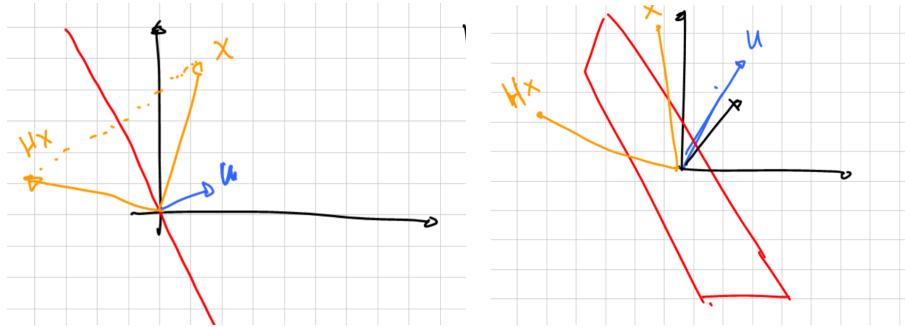


FIGURE 6.2: On the left in \mathbb{R}^2 , on the right in \mathbb{R}^3 .

$y = H^T x$, actually $H^T = H^{-1} = H$ Let's map x to y :

$$U = x - y = \begin{pmatrix} x_1 \\ x_2 \\ \vdots \\ x_n \end{pmatrix} - \begin{pmatrix} s \\ 0 \\ \vdots \\ 0 \end{pmatrix} = \begin{pmatrix} x_1 - s \\ x_2 \\ \vdots \\ x_n \end{pmatrix}$$

where $s = \|x\|$

Matlab syntax for functions

`function[v,s] = householder_vector(x)`, where v and s are the returned values and x is the argument.

Lecture 7

26th of October 2018 — F. Poloni

7.1 How to construct a QR factorization

In the previous lecture we introduced the QR factorization and we defined what an Householder reflector is.

At the end of the lecture we gave a first MatLab implementation of `householder_vector`:

ALGORITHM 7.1.1 Householder vector Matlab implementation.

```
1      function[v,s] = householdervector(x)
2          s = norm(x);
3          v = x;
4          v(1) = v(1) - s;
5          v = v / norm(v);
```

What's the problem of this algorithm? That the subtraction may create a problem with machine numbers, if s and $\|x\|$ are very close. If we take $\|x\| = -s$ the subtraction becomes an addition, and everything works well.

In the end, we would like to obtain this behaviour for every possible value for x and s , so line 2 may be modified as `s = - sign(x(1)) * norm(x)`.

Example 7.1.1. Given $A = \begin{pmatrix} \times & \times & \times & \times & \times \\ \times & \times & \times & \times & \times \\ \times & \times & \times & \times & \times \\ \times & \times & \times & \times & \times \\ \times & \times & \times & \times & \times \end{pmatrix} \in \mathcal{M}(m, m, \mathbb{R})$, where $m = 5$, we would like to calculate the QR factorization of A .



FIGURE 7.1: If x is oriented as in the plot it's better if we choose $-\|x\|e_1$ verse, since it's opposite to x .

STEP 1 : construct a Householder matrix that sends $A(:, 1)$ (first column of A)

to a multiple of e_1 . Then we have $H_1 A = \begin{pmatrix} \times & \times & \times & \times & \times \\ 0 & \times & \times & \times & \times \\ 0 & \times & \times & \times & \times \\ 0 & \times & \times & \times & \times \\ 0 & \times & \times & \times & \times \end{pmatrix}$

STEP 2 : take $H_2 \in \mathcal{M}(m-1, m-1, \mathbb{R})$ such that $H_2 A(2 : end, 2) = \begin{pmatrix} \times \\ 0 \\ 0 \\ 0 \end{pmatrix}$ and compute:

$$\begin{aligned} \begin{pmatrix} 1 & 0 & 0 & 0 & 0 \\ 0 & & & & \\ 0 & & H_2 & & \\ 0 & & & & \\ 0 & & & & \end{pmatrix} \cdot H_1 A &= \begin{pmatrix} 1 & 0 & 0 & 0 & 0 \\ 0 & & & & \\ 0 & & H_2 & & \\ 0 & & & & \\ 0 & & & & \end{pmatrix} \cdot \begin{pmatrix} \times & \times & \times & \times & \times \\ 0 & \times & \times & \times & \times \\ 0 & \times & \times & \times & \times \\ 0 & \times & \times & \times & \times \\ 0 & \times & \times & \times & \times \end{pmatrix} \\ &= \begin{pmatrix} \times & \times & \times & \times & \times \\ 0 & \times & \times & \times & \times \\ 0 & 0 & \times & \times & \times \\ 0 & 0 & \times & \times & \times \\ 0 & 0 & \times & \times & \times \end{pmatrix} \end{aligned} \tag{7.1.1}$$

And we denote $Q_2 = \begin{pmatrix} I_{1 \times 1} & 0 \\ 0 & H_2 \end{pmatrix}$, $Q_1 = H_1$;

STEP 3 : take $H_3 \in \mathcal{M}(m-2, m-2, \mathbb{R})$ such that $H_3 A(3 : end, 3) = \begin{pmatrix} \times \\ 0 \\ 0 \end{pmatrix}$ and

we compute:

$$\begin{aligned}
Q_3 \cdot (Q_2 Q_1 A) &= \begin{pmatrix} 1 & 0 & 0 & 0 & 0 \\ 0 & 1 & 0 & 0 & 0 \\ 0 & 0 & & & \\ 0 & 0 & & H_3 & \\ 0 & 0 & & & \end{pmatrix} \cdot \begin{pmatrix} \times & \times & \times & \times & \times \\ 0 & \times & \times & \times & \times \\ 0 & 0 & \times & \times & \times \\ 0 & 0 & \times & \times & \times \\ 0 & 0 & \times & \times & \times \end{pmatrix} \\
&= \begin{pmatrix} \times & \times & \times & \times & \times \\ 0 & \times & \times & \times & \times \\ 0 & 0 & \times & \times & \times \\ 0 & 0 & 0 & \times & \times \\ 0 & 0 & 0 & \times & \times \end{pmatrix}
\end{aligned} \tag{7.1.2}$$

$$\text{So, } Q_3 = \begin{pmatrix} I_{2 \times 2} & 0 \\ 0 & H_3 \end{pmatrix};$$

STEP 4 : take $H_4 \in \mathcal{M}(m-3, m-3, \mathbb{R})$ such that $H_4 A(4 : \text{end}, 4) = \begin{pmatrix} \times \\ 0 \end{pmatrix}$ and
we compute:

$$\begin{aligned}
Q_4 \cdot (Q_3 Q_2 Q_1 A) &= \begin{pmatrix} 1 & 0 & 0 & 0 & 0 \\ 0 & 1 & 0 & 0 & 0 \\ 0 & 0 & 1 & & \\ 0 & 0 & & H_4 & \\ 0 & 0 & & & \end{pmatrix} \cdot \begin{pmatrix} \times & \times & \times & \times & \times \\ 0 & \times & \times & \times & \times \\ 0 & 0 & \times & \times & \times \\ 0 & 0 & 0 & \times & \times \\ 0 & 0 & 0 & \times & \times \end{pmatrix} \\
&= \begin{pmatrix} \times & \times & \times & \times & \times \\ 0 & \times & \times & \times & \times \\ 0 & 0 & \times & \times & \times \\ 0 & 0 & 0 & \times & \times \\ 0 & 0 & 0 & 0 & \times \end{pmatrix}
\end{aligned} \tag{7.1.3}$$

$$\text{Where, } Q_4 = \begin{pmatrix} I_{3 \times 3} & 0 \\ 0 & H_4 \end{pmatrix}.$$

In the end, since Q_i is an orthogonal matrix and the product of orthogonal matrices is orthogonal, $Q_1 \cdot Q_2 \cdot Q_3 \cdot Q_4 A = T$, which is an upper triangular matrix.

Theorem 7.1.1 (Product of block matrices). Let $I \in \mathcal{M}(k, k, \mathbb{R})$, let $H_i \in \mathcal{M}(m-k, m-k, \mathbb{R})$ and let $B_i \in \mathcal{M}(k, k, \mathbb{R})$, $C_i \in \mathcal{M}(k, m-k, \mathbb{R})$ and $A_i \in \mathcal{M}(m-k, m-k, \mathbb{R})$, then the product between the two following block matrices is exactly the one showed below.

$$\begin{pmatrix} I & 0 \\ 0 & H_i \end{pmatrix} \cdot \begin{pmatrix} B_i & C_i \\ 0 & A_i \end{pmatrix} = \begin{pmatrix} B_i I & C_i \\ 0 & H_i \cdot A_i \end{pmatrix}$$

Proof. It's trivial computation, using the definition of matrix product. \square

7.1.1 Matlab implementation

ALGORITHM 7.1.2 First implementation of QR factorization.

```

1 function [Q, R] = myqr(A)
2 [m, n] = size(A);
3 Q = eye(m);
4 for j = 1:n
5     v = householder_vector(A(j:end, j));
6     H = eye(length(v)) - 2*v*v';
7     A(j:end,j:end) = H * A(j:end,j:end);
8     Q(:, j:end) = Q(:, j:end) * H;
9 end
10 R = A;
```

Fact 7.1.2. *The cost of this implementation when A is a square matrix is $O(n^3 + (n-1)^3 + \dots + 1^3)$. If A is a rectangular matrix, then the computational complexity is $O(m \cdot n^2 + (m-1) \cdot (n-1)^2 + \dots + (m-n+1)^3)$.*

Proof. Line 7 does a matrix product between matrices of size $n, n-1, \dots, 1$, so the resulting cost is $O(m \cdot n^2 + (m-1) \cdot (n-1)^2 + \dots + (m-n+1)^3)$. \square

We may design a faster algorithm, since $HA_j = A_j - 2v(v^T A_j)$.

ALGORITHM 7.1.3 More efficient implementation of QR factorization.

```

1 function [Q, A] = myqr(A)
2 [m, n] = size(A);
3 Q = eye(m);
4 for j = 1:n-1
5     [v, s] = householder_vector(A(j:end, j));
6     A(j, j) = s; A(j+1:end, j) = 0;
7     A(j:end, j+1:end) = A(j:end, j+1:end) - ...
8         2*v*(v'*A(j:end, j+1:end));
9     Q(:, j:end) = Q(:, j:end) - Q(:, j:end)*v*2*v';
10 end
```

Let's suppose that A is square matrix, partitioned as: $A = \begin{pmatrix} A_1 & A_2 \\ Q_1 & Q_2 \end{pmatrix} \cdot \begin{pmatrix} R_{1,1} & R_{1,2} \\ 0 & R_{2,2} \end{pmatrix}$ Then we can recover the factorization of A_1 from the factorization of A , since $A_1 = Q \cdot R_{11}$.

Fact 7.1.3 (Thin QR factorization). *we may replace $Q \in \mathcal{M}(m, m, \mathbb{R})$ and $R \in \mathcal{M}(m, m, \mathbb{R})$ with $Q_1 \in \mathcal{M}(m, n, \mathbb{R})$ and $R_1 \in \mathcal{M}(n, n, \mathbb{R})$ and the same factorization holds: $A = QR = Q_1 R_1$. This is called **thin QR factorization**.*

Proof. $A_1 \in \mathcal{M}(m, n, \mathbb{R})$, $A = \begin{pmatrix} Q_1 & Q_2 \end{pmatrix} \cdot \begin{pmatrix} R_1 \\ 0 \end{pmatrix} = Q \cdot R = \begin{pmatrix} Q_1 \end{pmatrix} \cdot (R_1) + (Q_2)(0) = Q_1 \cdot R_1$ \square

In order to save space we may work in the following way:

$$Q \cdot B = \begin{pmatrix} 1 & \times & \cdots & \times \\ 0 & & & \\ \vdots & & I - 2V_1 V_1^T & \\ 0 & & & \\ 0 & & & \end{pmatrix} \cdot \begin{pmatrix} 1 & \times & \times & \cdots & \times \\ 0 & 1 & \times & \cdots & \times \\ \vdots & 0 & & I - 2V_2 V_2^T & \\ 0 & \vdots & & & \\ 0 & 0 & & & \end{pmatrix} \cdot \dots \cdot \begin{pmatrix} 1 & \times & \times & \cdots & \times \\ 0 & 1 & \times & \cdots & \times \\ \vdots & 0 & 1 & & \\ 0 & \vdots & 0 & I - 2V_n V_n^T & \\ 0 & 0 & 0 & & \end{pmatrix} \cdot B.$$

Fun fact

There are some libraries that store the v_i vectors in the lower part of matrix R which is upper triangular and has only zeros below the main diagonal.

7.2 How to use the thin QR factorization to solve a least squares problem

We would like to solve $\|Ax - b\|$ $\forall A \in \mathcal{M}(m, n, \mathbb{R})$ and $\forall B \in \mathbb{R}^n$ where $m > n$ (a.k.a A is a tall, thin matrix), through the QR factorization. We would like to solve $\min \|Ax - b\|$ through the QR factorization.

We may write first the QR factorization of A , so $\forall A \in \mathcal{M}(m, n, \mathbb{R})$, $\exists Q \in \mathcal{M}(m, m, \mathbb{R})$, $\exists R \in \mathcal{M}(m, n, \mathbb{R})$ such that $A = QR$, where $Q = \begin{pmatrix} Q_1 & Q_2 \end{pmatrix}$ and $R = \begin{pmatrix} R_1 \\ 0 \end{pmatrix}$.

Then,

$$\begin{aligned}
\|Ax - \mathbf{b}\| &= \|Q^T(Ax - \mathbf{b})\| = \|Q^T Q R \mathbf{x} - Q^T \mathbf{b}\| \\
&= \|R \mathbf{x} - Q^T \mathbf{b}\| = \left\| \begin{pmatrix} R_1 \\ 0 \end{pmatrix} \mathbf{x} - \begin{pmatrix} Q_1^T \\ Q_2^T \end{pmatrix} \mathbf{b} \right\| \\
&= \left\| \begin{pmatrix} R_1 \mathbf{x} - Q_1^T \mathbf{b} \\ Q_2^T \mathbf{b} \end{pmatrix} \right\|
\end{aligned} \tag{7.2.1}$$

How can we pick \mathbf{x} to minimize the norm of $A\mathbf{x} - \mathbf{b}$?

Can we choose x such that $R_1 \mathbf{x} - Q_1^T \mathbf{b} = 0$? Yes, we can, since this is a linear square system, so $\mathbf{x} = R_1^{-1} Q_1^T \mathbf{b}$

$$\|Ax - \mathbf{b}\| = \|Q_2^T \mathbf{b}\|$$

We used the fact that R_1 is invertible, but is it always true that R_1 is invertible?

Lemma 7.2.1. R_1 is invertible $\Leftrightarrow A$ has full column rank.

Proof. A has full column rank $\Leftrightarrow A^T A$ is positive definite $\Leftrightarrow A^T A$ is positive semidefinite and invertible, but $A^T A$ is positive semidefinite, so we only need to prove its invertibility.

Let's compute $Q R^T Q R = R^T Q^T Q R = R^T R = (R_1^T \ 0) \cdot \begin{pmatrix} R_1 \\ 0 \end{pmatrix} = R_1^T \cdot R_1$.

So, $A^T A$ is invertible $\Leftrightarrow R_1$ is invertible. \square

Note

$R_1^T R_1$ is the Cholesky factorization of $A^T A$.

The computational complexity is asymptotically equal to the one of computing the QR factorization, since the other operations are cheaper (the product $Q_1^T \mathbf{b}$ costs $O(mn)$ and solving the triangular linear system by back-substitution costs $O(n^2)$).

Lecture 8

7th of November 2018 —
F.Poloni

8.1 Least squares problem with SVD

💡 Do you recall?

Tall thin SVD: A matrix A can be written as $A = USV^T$, where U is orthogonal, S is a diagonal matrix and V is orthogonal as well. In the case of a tall, thin A the decomposition has the following shape:

$$\begin{pmatrix} & & & \\ & & & \\ U^1 & U^2 & \dots & U^m \\ & & & \end{pmatrix} \cdot \Sigma \cdot \begin{pmatrix} V^1 & V^2 & \dots & V^n \end{pmatrix}^T$$

where

$$\Sigma = \begin{pmatrix} \sigma_1 & & & \\ & \sigma_2 & & \\ & & \ddots & \\ & & & \sigma_n \end{pmatrix}$$

And if we denote U_1 the matrix obtained as the first n columns of U we have the tall, thin SVD: $A = U_1 \cdot \Sigma \cdot V^T$

We would like to see how we can solve a least squares problem through *SVD*:

$$\begin{aligned}
\|Ax - b\| &= \|USV^T x - b\| \quad \leftarrow \text{Def. of SVD} \\
&= \|U^T(USV^T x - b)\| \quad \leftarrow U^T \text{ is orthogonal} \\
&= \|SV^T x - U^T b\| \quad \leftarrow \text{Distributivity + orthogonality} \\
&= \|Sy - U^T b\| \quad \leftarrow y = V^T x \\
&= \left\| \begin{pmatrix} \sigma_1 & & & \\ & \sigma_2 & & \\ & & \ddots & \\ & & & \sigma_n \\ & & & 0 \end{pmatrix} \begin{pmatrix} y_1 \\ y_2 \\ \vdots \\ y_n \end{pmatrix} - \begin{pmatrix} U^{1T} b \\ U^{2T} b \\ \vdots \\ U^{nT} b \\ U^{n+1T} b \\ \vdots \\ U^{mT} b \end{pmatrix} \right\| \quad (8.1.1) \\
&= \left\| \begin{pmatrix} \sigma_1 y_1 - U^{1T} b \\ \sigma_2 y_2 - U^{2T} b \\ \vdots \\ \sigma_n y_n - U^{nT} b \\ \sigma_{n+1} y_{n+1} - U^{n+1T} b \\ \vdots \\ \sigma_m y_m - U^{mT} b \end{pmatrix} \right\|
\end{aligned}$$

Where the first n rows may be assigned to 0 iff $y_i = -\frac{U^{iT} b}{\sigma_i}$ (if $\sigma_i \neq 0 \forall i$), while the latter $m - n$ do not depend on y . This process produces a solution y , but

the variable change may be inverted, so

$$\begin{aligned}
x &= Vy \quad \leftarrow \text{Orthogonality of } V \\
&= V^1 y_1 + V^2 y_2 + \cdots + V^n y_n \\
&= V^1 \frac{1}{\sigma_1} U^{1T} b + V^2 \frac{1}{\sigma_2} U^{2T} b + \cdots + V^n \frac{1}{\sigma_n} U^{nT} b \\
&= V \cdot \begin{pmatrix} \frac{1}{\sigma_1} & & & \\ & \frac{1}{\sigma_2} & & \\ & & \ddots & \\ & & & \frac{1}{\sigma_n} \end{pmatrix} \cdot U^T \cdot b \tag{8.1.2} \\
&= V \cdot \begin{pmatrix} \frac{1}{\sigma_1} & & & \\ & \frac{1}{\sigma_2} & & \\ & & \ddots & \\ & & & \frac{1}{\sigma_n} \end{pmatrix} \cdot U_1^T \cdot b
\end{aligned}$$

Which depends only on the tall, thin SVD.

Fact 8.1.1. *The σ_i are different from 0 iff A has full column rank.*

$$\begin{aligned}
&\text{Proof. } A \text{ has full column rank} \\
&\Updownarrow \\
&A^T A \text{ is invertible} \\
&\Updownarrow \\
&(USV^T)^T (USV^T) = VS^T U^T USV^T = V \cdot \begin{pmatrix} \sigma_1^2 & & & \\ & \sigma_2^2 & & \\ & & \ddots & \\ & & & \sigma_n^2 \end{pmatrix} \cdot V^T \text{ is} \\
&\text{invertible} \\
&\Updownarrow \\
&\forall i \ \sigma_i \neq 0
\end{aligned}$$

□

Observation 8.1.1. *This lemma also proves that the factorization is also a QR factorization.*

Note on Matlab syntax

`svd(A, 0)` and `qr(A, 0)` express that we are only interested in the parts of the factorization without zeros, in case of a tall, thin matrix A .

We may observe that the computational complexity is $O(15 \cdot n^3)$ for square matrices, while it's $O(m \cdot n^2)$ in the tall, thin case.

8.1.1 Behaviour in case of zeros as singular values

What happens when there are some zeros as singular values?

💡 Do you recall?

We may recall that the singular values are ordered on the diagonal in decreasing order (the largest in top left position). From this assumption, we may say that if there are some $\sigma_i = 0$ then they are in the bottom right part of the matrix.

$$\sigma_1 \geq \sigma_2 \geq \cdots \geq \sigma_r > \sigma_{r+1} = \cdots = \sigma_n = 0$$

$$\left\| \begin{pmatrix} \sigma_1 y_1 - U^{1T} b \\ \sigma_2 y_2 - U^{2T} b \\ \vdots \\ \sigma_n y_n - U^{nT} b \\ 0 \cdot y_{n+1} - U^{n+1T} b \\ \vdots \\ 0 \cdot y_m - U^{mT} b \end{pmatrix} \right\|.$$

We also recall the following $\|Ax - b\| =$

No matter what value we choose for $y_{r+1} \cdots y_n$, the value doesn't change since

$$\begin{pmatrix} -\frac{U^{1T} b}{\sigma_1} \\ -\frac{U^{2T} b}{\sigma_2} \\ \vdots \\ -\frac{U^{rT} b}{\sigma_r} \\ * \end{pmatrix}$$

it's multiplied by 0. Therefore we get infinite solutions of the form $y =$

We would like to make the solution unique, so we can modify the problem:

- taking the value that minimize the norm:

$$\min_{x \in \arg \min(\|Ax - b\|)} \|x\|. \quad (\text{P2})$$

Note that $\|x\| = \|y\|$, because $x = Vy$. It follows from the expression of y that the choice that minimizes its norm is $y_{r+1} = \cdots = y_n = 0$.

•

The solution of $P2$ is given by

$$V y = \begin{pmatrix} V^1 & V^2 & \cdots & V^n \end{pmatrix} \cdot \begin{pmatrix} -\frac{U^{1T} b}{\sigma_1} \\ -\frac{U^{2T} b}{\sigma_2} \\ \vdots \\ -\frac{U^{rT} b}{\sigma_r} \\ 0 \end{pmatrix} = V^{1T} \frac{1}{\sigma_1} U^{1T} b + \cdots + V^{rT} \frac{1}{\sigma_r} U^{rT} b$$

What happens when working with machine precision? Let's make an example where $r = n - 1$, so only the last singular value is 0. If the check of $\sigma_n = 0$ fails, $\frac{1}{\sigma_n}$ becomes very big. A way to circumvent this problem is to find the linear dependencies between the columns, so that the algorithm works correctly.

8.2 Truncated SVD

In many real world setups first singular components correspond to the most prominent features of the dataset, while the smallest ones are fine details and noise. Note, though, that in the sum $\sum_{i=1}^n V^i \frac{U^{iT} b}{\sigma_i}$ the small singular values may have a large impact, because σ_i is in the denominator.

We can modify the solution to cope with real world data problems:

$$x = \sum_{i=1}^n V^i \frac{U^{iT} b}{\sigma_i} \longrightarrow x_{trunc} = \sum_{i=1}^k V^i \frac{U^{iT} b}{\sigma_i}$$

for a certain k , ignoring small singular values.

Another way to modify the problem is the following.

8.3 Tikhonov regularization / ridge regression

The Tikhonov regularization is a smoother version of truncated SVD.

$$x_{Tik} = \arg \min_{x \in \mathbb{R}^n} \|Ax - b\|^2 + \alpha^2 \|x\|^2$$

Fact 8.3.1. *The Tikhonov regularization is equivalent to*

$$x_{Tik} = \arg \min_{x \in \mathbb{R}^n} \left\| \begin{pmatrix} A \\ \alpha I \end{pmatrix} x - \begin{pmatrix} b \\ 0 \end{pmatrix} \right\|^2 \quad (8.3.1)$$

We show that the two objective functions coincide.

Proof.

$$\begin{aligned}
\left\| \begin{pmatrix} A \\ \alpha I \end{pmatrix} x - \begin{pmatrix} b \\ 0 \end{pmatrix} \right\|^2 &= \left\| \begin{pmatrix} Ax \\ \alpha x \end{pmatrix} - \begin{pmatrix} b \\ 0 \end{pmatrix} \right\|^2 \\
&= \left\| \begin{pmatrix} Ax - b \\ \alpha x \end{pmatrix} \right\|^2 \\
&= \|Ax - b\|^2 + \|\alpha x\|^2 \\
&= \|Ax - b\|^2 + \alpha \cdot \|x\|^2
\end{aligned} \tag{8.3.2}$$

□

Fact 8.3.2. *The solution of the Tikhonov regularization is given by the formula $x_{Tik} = (A^T A + \alpha^2 I)^{-1} A^T b$.*

Proof. We start by writing the explicit solution of Equation (8.3.1) using the pseudoinverse

$$\begin{aligned}
\begin{pmatrix} A \\ \alpha I \end{pmatrix}^+ \begin{pmatrix} b \\ 0 \end{pmatrix} &= \left(\begin{pmatrix} A \\ \alpha I \end{pmatrix}^T \begin{pmatrix} A \\ \alpha I \end{pmatrix} \right)^{-1} \begin{pmatrix} A \\ \alpha I \end{pmatrix}^T \begin{pmatrix} b \\ 0 \end{pmatrix} \\
&= \left((A^T \quad \alpha I) \begin{pmatrix} A \\ \alpha I \end{pmatrix} \right)^{-1} (A^T \quad \alpha I) \begin{pmatrix} b \\ 0 \end{pmatrix} \\
&= (A^T A + \alpha^2 I)^{-1} A^T b.
\end{aligned} \tag{8.3.3}$$

□

Fact 8.3.3. *We can observe that $(A^T A + \alpha^2 I)$ is positive definite.*

Proof. $z^T \cdot (A^T A + \alpha^2 I) \cdot z = z^T A^T A z + \alpha^2 z^T z \stackrel{(1)}{=} \alpha^2 z^T z = \alpha^2 \|z\|^2 > 0$. The equality (1) is obtained because $z^T A^T A z \geq 0$, since $A^T A$ is positive semidefinite.

□

Exercise 8.3.1. *Show using the SVD of A that the Tikhonov / Ridge solution can be written as*

$$x = \sum_{i=1}^n V^i \frac{\sigma_i}{\sigma_i^2 + \alpha^2} U^{i^T} b.$$

When $\sigma_i \gg \alpha$, $\frac{\sigma_i}{\sigma_i^2 + \alpha^2} \approx \frac{1}{\sigma_i}$: similar to the ‘true’ solution.

When $\sigma_i \ll \alpha$, $\frac{\sigma_i}{\sigma_i^2 + \alpha^2} \approx \frac{\sigma_i}{\alpha^2} \approx 0$: approximately ignoring small singular values.

Per casa: fare esercizio.

How can we choose k ? We don’t know.

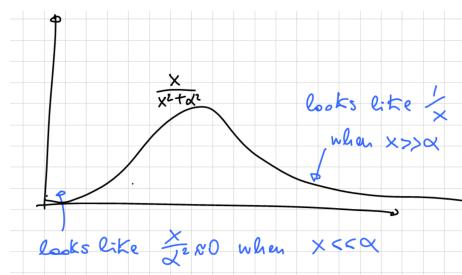


FIGURE 8.1: Here is the shape of the formula for the singular values.

Lecture 9

9th of November 2018 — F. Poloni

9.1 Conditioning

Two lectures ago we introduced the QR factorization to solve least squares problems and we noticed that it has a computational complexity which is much worse than the normal equations method.

Why did we introduce the QR factorization to solve the least squares problem, then? Although it's more complex computationally speaking, it's much better than the normal equation for what concerns accuracy. Let's see why through an example:

Example 9.1.1. Let $A \in \mathcal{M}(4, 3, \mathbb{R})$ s.t.

$$A = \begin{pmatrix} 1 & 1 & 2 \\ 1 & 2 & 3 \\ 3 & 1 & 4 \\ 1 & 2 & 3 + 10^{-8} \end{pmatrix}$$

In this case the normal equations method doesn't even find the order of magnitude correctly.

At this point we may introduce the problem of **sensitivity**:

Definition 9.1.1 (Sensitivity). We call **sensitivity** the measure of how much the output of a problem changes when we perturb its input.

As an example, let $f(x, y) = x + 2y$. If we perturb the second parameter of f as follows $\tilde{y} = y + \delta$, we can compute the variation in the output value of the function v as

$$\mathbf{v} = f(x, \tilde{y}) - f(x, y) = x + 2(y + \delta) - (x + 2y) = 2\delta.$$

A good example of this behaviour is the temperature of water coming from the shower: in particular, when we rotate little the knob the water becomes too cold or too hot very fast. This function is plotted in Figure 9.1.

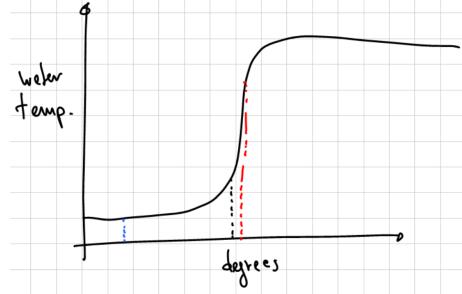


FIGURE 9.1: Geometric idea of temperature of the water in the shower

Definition 9.1.2 (Absolute condition measure). *The absolute condition number of a function f is the maximum possible output change / input change ratio in the limit for a small change of the input.*

$$\kappa_{abs}(f, x) = \lim_{\varepsilon \rightarrow 0} \sup_{|\tilde{x}-x| \leq \varepsilon} \frac{|f(\tilde{x}) - f(x)|}{|\tilde{x} - x|}.$$

We would like to focus on what this definition means.

Why are we interested in the limit of a very small change?

If we zoom-in a continuous function it gets basically linear (key idea of derivative) and then the ratio between the difference on the outputs and the one of the inputs is approximatively the derivative, as shown in Figure 9.2.

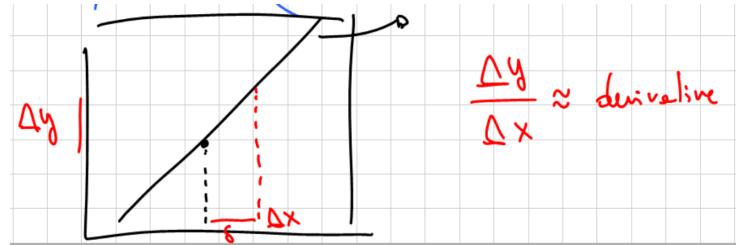


FIGURE 9.2: Geometric idea behind the derivative, as “zoom” of the function in a certain point.

We take a point x and we consider a ball of radius ε and we compute the change in the output over the change in the input, then we take the maximum.

Example 9.1.2. Let $f(x) = x^2$.

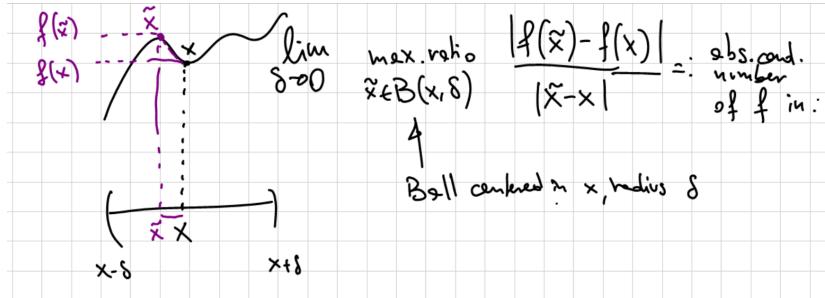


FIGURE 9.3: Geometric idea behind absolute condition number.

If we perturb the input x to $x + \delta$, then $f(\tilde{x}) = (x + \delta)^2 = x^2 + 2x\delta + \delta^2$, then we obtain the ratio

$$r = \frac{|f(\tilde{x}) - f(x)|}{|\tilde{x} - x|} = \frac{|2x\delta + \delta^2|}{|\delta|} = |2x + \delta|$$

If we denote ε the radius of the ball, we obtain the following

$$\max_{\substack{|\delta| < |\varepsilon| \\ \tilde{x} \in B(x, \varepsilon)}} r = |2x| + |\varepsilon|$$

then

$$\lim_{\varepsilon \rightarrow 0} \max_{|\delta| < |\varepsilon|} r = \lim_{\varepsilon \rightarrow 0} |2x| + |\varepsilon| = |2x|$$

Example 9.1.3. It's more interesting to see a multivariate function:

Let $f(x) = x^T Q x$, for instance for $x \in \mathbb{R}^2$ so that we can plot its graph in \mathbb{R}^3 .

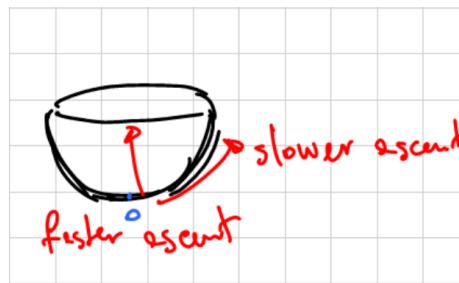


FIGURE 9.4: Paraboloid

We shall take a general example where the cross-section are ellipses, so that there is a direction of faster and slower ascent; this is not just a circular "cup"

seen in perspective. Note that these directions of faster ascent and lower ascent correspond to the eigenvectors of the matrix Q .

In this case the absolute condition number is $\lim_{\varepsilon \rightarrow 0} \max_{\tilde{x} \in B(x, \varepsilon)} \frac{\|f(\tilde{x}) - f(x)\|}{\|\tilde{x} - x\|}$, and one can see that the output/input ratio varies with the direction in which \tilde{x} is, so we have to take a maximum in the whole ball $B(x, \varepsilon)$.

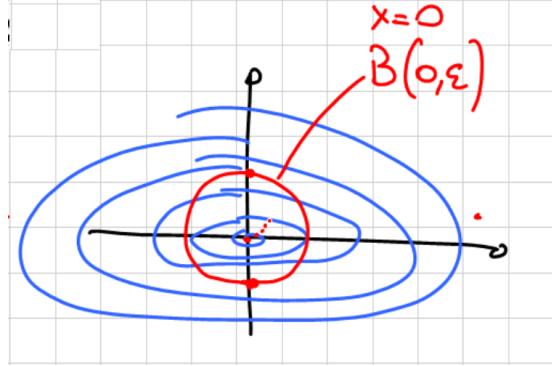


FIGURE 9.5: Level curves of a quadratic function (“seen from above”).

At this point, an observation is mandatory: **any** absolute measure doesn't take into account the values of the function in other points, so we want to define the following

Definition 9.1.3 (Relative error). *The relative error of an approximation \tilde{x} to a quantity x is $\frac{\|\tilde{x} - x\|}{\|x\|}$.*

Here are some examples of good and bad accuracy:

- $\frac{|\tilde{x} - x|}{|x|} \approx 1$: **very bad** accuracy; it's just a number with the same order of magnitude.
- $\frac{|\tilde{x} - x|}{|x|} \approx 10^{-3}$: about 3 correct significant digits.
- $\frac{|\tilde{x} - x|}{|x|} \approx 10^{-16}$: about 16 correct digits; we **can't do better** typically (with double precision arithmetic).

Definition 9.1.4 (Relative condition number). *The relative condition number of a function f is defined as*

$$\kappa_{rel}(f, \mathbf{x}) = \lim_{\delta \rightarrow 0} \sup_{\frac{\|\tilde{\mathbf{x}} - \mathbf{x}\|}{\|\mathbf{x}\|} \leq \delta} \frac{\frac{\|f(\tilde{\mathbf{x}}) - f(\mathbf{x})\|}{\|f(\mathbf{x})\|}}{\frac{\|\tilde{\mathbf{x}} - \mathbf{x}\|}{\|\mathbf{x}\|}} = \kappa_{abs}(f, \mathbf{x}) \frac{\|\mathbf{x}\|}{\|f(\mathbf{x})\|},$$

i.e., we replace the absolute error $\|\tilde{\mathbf{x}} - \mathbf{x}\|$ with the relative error.

9.1.1 Conditioning of linear systems

At this point we would like to compute the condition number of solving a linear system, i.e., the condition number of the function $f(A, b) = A^{-1}b$, perturbing the inputs A and b , one at a time.

PERTURBING b We want to compute the limit of the relative error $\frac{\|f(A, \tilde{b}) - f(A, b)\|}{\|f(A, b)\|}$,

so we set $x = A^{-1}b$ and $\tilde{x} = A^{-1}\tilde{b}$, and we estimate *output error* = $\frac{\|\tilde{x} - x\|}{\|x\|} = ?$

1.

$$\begin{aligned} \|\tilde{x} - x\| &= \|A^{-1}\tilde{b} - A^{-1}b\| \\ &= \|A^{-1}(\tilde{b} - b)\| \\ &\leq \|A^{-1}\| \cdot \|\tilde{b} - b\| \end{aligned} \quad (9.1.1)$$

2. Since $\|b\| = \|Ax\| \leq \|A\| \cdot \|x\|$ we have $\frac{\|\tilde{x} - x\|}{\|x\|} \leq \|A^{-1}\| \cdot \|A\| \cdot \frac{\|\tilde{b} - b\|}{\|b\|}$

In the end, since *input error* = $\frac{\|\tilde{b} - b\|}{\|b\|}$ we obtain

$$\kappa_{rel}(f, x) = \lim_{\varepsilon \rightarrow 0} \frac{\text{output error}}{\text{input error}} \leq \lim_{\varepsilon \rightarrow 0} \|A^{-1}\| \cdot \|A\| = \|A^{-1}\| \cdot \|A\|$$

We denote $\kappa(A) = \|A^{-1}\| \cdot \|A\|$ the **condition number of A** ;

PERTURBING A Given $Ax = b$ we obtain $(A + \Delta_A) \cdot (x + \Delta_x) = b$, where $\tilde{A} = A + \Delta_A$ and $\tilde{x} = x + \Delta_x$. Then we can expand as follows

$$A\tilde{x} + \Delta_A \cdot x + A \cdot \Delta_x + \Delta_A \cdot \Delta_x = \tilde{b}$$

We can stop taking into account $\Delta_A \cdot \Delta_x$, since it's a sort of second order term ($\Delta_A \cdot \Delta_x = o(\|\Delta_A\| \cdot \|\Delta_x\|)$), so we get the following

$$\Delta_A \cdot x + A \cdot \Delta_x = 0$$

$$\Delta_x = -A^{-1} \cdot \Delta_A \cdot x$$

then $\|\Delta_x\| \leq \|A^{-1}\| \cdot \|\Delta_A\| \cdot \|x\|$, which implies $\frac{\|\Delta_x\|}{\|x\|} \leq \|A^{-1}\| \cdot \frac{\|\Delta_A\|}{\|A\|}$.

We obtain that *relative output error* $\leq \kappa(A) \cdot \text{relative input error}$.

We only proved an inequality, but it turns out that it is tight: for every A and b there is a possible choice of the perturbation \tilde{x} that attains equality.

In the end, in both cases, the error in the output is the error in the input (namely b or A) times the condition number.

9.2 Condition number, SVD, and distance to singularity

Fact 9.2.1. $\kappa(A) = \frac{\sigma_1}{\sigma_n}$, i.e., $\kappa(A)$ is the ratio between the smallest and the largest singular value.

So we can say that if a matrix is close to a singular matrix, then its condition number is going to be large.

Proof. Let $A = USV^T$, then $\|A\| = \|U \cdot S \cdot V^T\| = \|S\| = \sigma_1$, since

$$S = \begin{pmatrix} \sigma_1 & & & \\ & \sigma_2 & & \\ & & \ddots & \\ & & & \sigma_n \end{pmatrix} \text{ and } \sigma_1 \geq \sigma_2 \geq \cdots \geq \sigma_n.$$

It's also true that $\|A^{-1}\| = \|(USV)^{-1}\| = \|VS^{-1}U^T\| = \|S^{-1}\| = \sigma_n$, since

$$S^{-1} = \begin{pmatrix} \frac{1}{\sigma_1} & & & \\ & \frac{1}{\sigma_2} & & \\ & & \ddots & \\ & & & \frac{1}{\sigma_n} \end{pmatrix} \text{ and } \sigma_1 \geq \sigma_2 \geq \cdots \geq \sigma_n.$$

In the end $\kappa(A) = \frac{\|A\|}{\|A^{-1}\|} = \frac{\sigma_1}{\sigma_n}$

□

Fact 9.2.2. The relative distance between A and the closest singular matrix is $\frac{1}{\kappa(A)}$.

💡 Do you recall?

Eckart-Young theorem: the closest matrix to A that has rank $\leq n-1$ is

$$\hat{A} = U \begin{pmatrix} \sigma_1 & & & \\ & \sigma_2 & & \\ & & \ddots & \\ & & & \sigma_n \\ & & & 0 \end{pmatrix} \cdot V^T$$

Proof.

bau (9.2.1)

$$\begin{aligned}
\|A - \hat{A}\| &= \left\| U \cdot \left(\begin{pmatrix} \sigma_1 & & & \\ & \sigma_2 & & \\ & & \ddots & \\ & & & \sigma_n \end{pmatrix} - \begin{pmatrix} \sigma_1 & & & \\ & \sigma_2 & & \\ & & \ddots & \\ & & & \sigma_{n-1} & 0 \end{pmatrix} \right) \cdot V^T \right\| \\
&= \left\| U \cdot \begin{pmatrix} 0 & & & \\ & \ddots & & \\ & & 0 & \\ & & & \sigma_n \end{pmatrix} \cdot V^T \right\| \\
&= \sigma_n
\end{aligned} \tag{9.2.2}$$

Thus, $\|A - \hat{A}\| = \sigma_n$. We know already that $\|A\| = \sigma_1$, so we just need to take the ratio. \square

We analyzed the conditioning of linear systems, but the main problem we want to study in this course is least squares problem.

9.3 Conditioning of least squares problem

We need two quantities to be able to measure the conditioning of least squares problem:

★ $\kappa(A)$ Let $A \in \mathcal{M}(m, n, \mathbb{R})$, with $m > n$ (tall, thin A). We define $\kappa(A) = \frac{\sigma_1}{\sigma_n}$. We take this as a definition of $\kappa(A)$. Note that we cannot use the other definition $\kappa(A) = \|A\| \|A^{-1}\|$, since A^{-1} does not exist for a non-square A . However, one can verify that $\|A\| \cdot \|A^\dagger\| = \frac{\sigma_1}{\sigma_n} = \kappa(A)$, where A^\dagger is the pseudoinverse.

Observation 9.3.1. Note that $\frac{1}{\kappa(A)}$ is the relative distance to the closest \hat{A} without full column rank.

★ θ The second quantity needed is the angle between Ax and b , see Figure 9.6.
 $\theta = \arccos \frac{\|Ax\|}{\|b\|}$

Now we can express the theorem:

Theorem 9.3.1 (Trefethen, Bau, Theorem 18.1). Consider the linear least squares problem $\min \|Ax - b\|$, with $A \in \mathbb{R}^{m \times n}$ with full column rank. Its relative condition number with respect to the input b is

$$\kappa_{rel,b \rightarrow x} \leq \frac{\kappa(A)}{\cos \theta},$$

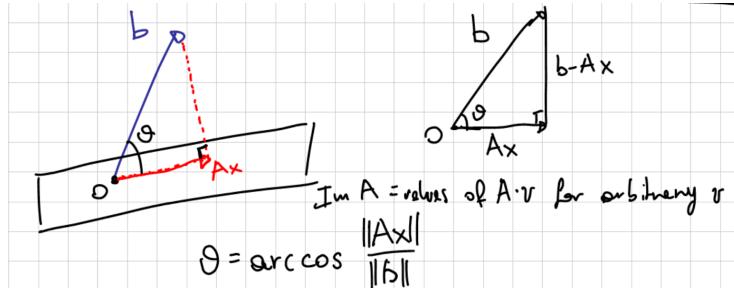


FIGURE 9.6: The triangle in the figure (the one whose cathetus are Ax and $b - Ax$) is a square triangle.

and with respect to A it is

$$\kappa_{rel,A \rightarrow x} \leq \kappa(A) + \kappa(A)^2 \tan \theta,$$

where θ is the angle such that $\cos \theta = \frac{\|Ax\|}{\|b\|}$.

At this point we have two condition numbers and they both depend on $\kappa(A)$ and θ .

Observation 9.3.2.

SPECIAL CASE 1: $\theta \approx 90^\circ$ We can see from the figure that a big change of b induces a small perturbation of Ax . No matter what the conditioning of A is, a small (relative) perturbation in b can change a large (relative) perturbation in x and Ax , see Figure 9.7.

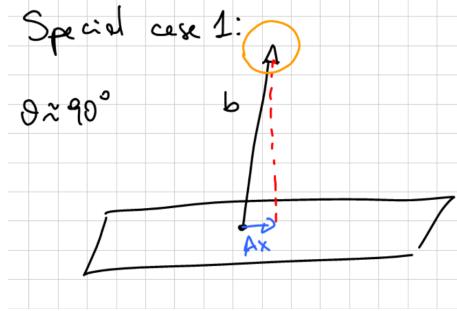


FIGURE 9.7: Special case 1

SPECIAL CASE 2: $\theta \approx 0^\circ$ When b is almost in plane with $Im(x)$. In this case $cond \approx \kappa(A)$, see Figure 9.8.

GENERAL CASE: θ FAR FROM 0° AND 90° In the more general case, $cond \approx \kappa(A)^2$.

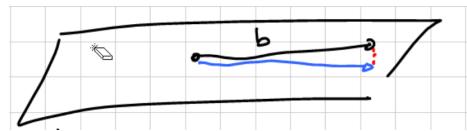


FIGURE 9.8: Special case 2.

Lecture 10

15th of November 2018 — F. Poloni

10.1 Stability of algorithms

In this lecture we will try to answer the question: “Is our algorithm (using floating point) going to compute a good approximation of the answer?”

It is related to sensitivity / conditioning but different. Depends on how we perform the computation.

💡 Do you recall?

Computers work with IEEE arithmetic and the basic idea is the following.
TL;DR: floating point numbers are numbers in base-2 scientific (exponential) notation.

`double` (64-bit numbers):

$$\pm 1.\underbrace{01001011101\dots 101}_{52 \text{ binary digits}} \cdot 2^{\pm \underbrace{101\dots 01}_{10 \text{ binary digits}}}.$$

We use 1 bit for the sign, 52 bits for the “mantissa” and 11 bits for the exponent and its sign.

Some of these combinations of bits are reserved for special numbers, e.g. `Inf` and `NaN`, `-0`.

This system is subject to approximation errors, exactly like the “usual” decimal arithmetic: for example, if we do $\frac{1}{3} = 0.33333\dots$ and if we do $\frac{1}{3} + \frac{1}{3} + \frac{1}{3} = 0.99999\dots \neq 1$.

Whenever we store a number on a computer we need to approximate it, using the “representable numbers”, as shown in Figure 10.1.

Not all numbers are exactly representable, take 0.1 (decimal). It’s a periodic

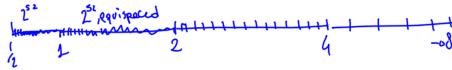


FIGURE 10.1: We have 2^{52} equispaced numbers between $\frac{1}{2}$ and 1 and between 1 and 2, and also 2^{52} between 2 and 4 and so on and so forth, so we have the same number of integers, although the space is enlarging.

number when written in binary, hence we can't represent it exactly as a machine number.

Definition 10.1.1 (Error bound). *For each $x \in \pm[10^{-308}, 10^{308}]$, there is an exactly representable number \tilde{x} such that $\frac{|\tilde{x}-x|}{|x|} \leq u$, with $u = 2^{-52} \approx 2 \cdot 10^{-16}$.*

Let us assume that we have the best possible algorithm that returns $\tilde{y} = f(\tilde{x})$ which is the best representation of $f(x)$, then

$$\begin{aligned} \frac{|\tilde{y}-y|}{|y|} &\leq \kappa_{rel}(f, x) \frac{|\tilde{x}-x|}{|x|} + o\left(\frac{|\tilde{x}-x|}{|x|}\right) \\ &\leq \kappa_{rel}(f, x)u + o(u). \end{aligned}$$

In practice we may ignore $o(u)$, since it's small o of 2^{-16} .

When we are approximating a result, we are approximating it up to a relative error of order of magnitude 10^{-16} .

$$a \oplus b = (a+b)(1+\delta), \quad |\delta| \leq u.$$

⇓

$$\frac{|(a \oplus b) - (a+b)|}{|a+b|} = \delta$$

We would like to compute the error on the function $f(a, b) = a^T b$.

$$a = \begin{pmatrix} a_1 \\ a_2 \\ a_3 \end{pmatrix}, b = \begin{pmatrix} b_1 \\ b_2 \\ b_3 \end{pmatrix}, \text{ so } f(a, b) = a_1 \cdot b_1 + a_2 \cdot b_2 + a_3 \cdot b_3$$

Denoting \odot the approximated product and \oplus the approximated sum on a computer, we can write the following:

$$\begin{aligned}
\text{computer result} &= a_1 \odot b_1 \oplus a_2 \odot b_2 \oplus a_3 \odot b_3 \\
&= \left[[a_1 b_1 \cdot (1 + \delta) + a_2 b_2 \cdot (1 + \delta_2)] \cdot (1 + \delta_3) + a_3 b_3 \cdot (1 + \delta_4) \right] \cdot (1 + \delta_5) \\
&= a_1 b_1 \cdot (1 + \delta_1) \cdot (1 + \delta_3) \cdot (1 + \delta_5) + a_2 b_2 \cdot (1 + \delta_2) \cdot (1 + \delta_3) \cdot (1 + \delta_5) \\
&\quad + a_3 b_3 \cdot (1 + \delta_4) \cdot (1 + \delta_5) \\
&= a_1 b_1 (1 + \delta_1 + \delta_3 + \delta_5 + O(u^2)) \\
&\quad + a_2 b_2 (1 + \delta_1 + \delta_3 + \delta_5 + O(u^2)) \\
&\quad + a_3 b_3 (1 + \delta_4 + \delta_5 + O(u^2)) \\
&\approx a_1 b_1 (1 + \delta_1 + \delta_3 + \delta_5) + a_2 b_2 (1 + \delta_1 + \delta_3 + \delta_5) + a_3 b_3 (1 + \delta_4 + \delta_5)
\end{aligned} \tag{10.1.1}$$

Where $O(u^2)$ comes from the summation of $\delta_i \delta_j$, for some i, j and allows us to do an approximation up to second order terms of precision.

The absolute error then is

$$\begin{aligned}
\text{Err}_a &= |a_1 b_1 (1 + \delta_1 + \delta_3 + \delta_5) + a_2 b_2 (1 + \delta_1 + \delta_3 + \delta_5) + a_3 b_3 (1 + \delta_4 + \delta_5) + a_1 b_1 + a_2 b_2 + a_3 b_3| \\
&\stackrel{(1)}{\leq} |a_1 b_1| \cdot 3u + |a_2 b_2| \cdot 3u + |a_3 b_3| \cdot 2u \\
&\leq (|a_1 b_1| + |a_2 b_2| + |a_3 b_3|) \cdot 3u
\end{aligned} \tag{10.1.2}$$

Where $\leq^{(1)}$ follows from the observation that $|\delta_i| \leq u$.

The result that we can obtain is weaker than we would have expected: if $a_i b_i \geq 0, \forall i = 1, 2, 3$, then

$$\frac{|\text{computer result} - (a_1 b_1 + a_2 b_2 + a_3 b_3)|}{a_1 b_1 + a_2 b_2 + a_3 b_3} \leq 3u,$$

which means that the algorithm is stable.

However, if $a_1 b_1, a_2 b_2$ and $a_3 b_3$ have different signs, then we can bound the error not with $a_1 b_1 + a_2 b_2 + a_3 b_3$, but only with $|a_1 b_1| + |a_2 b_2| + |a_3 b_3|$. This might be a lot larger than what we want to compute.

Example 10.1.1. Take $\varepsilon = 10^{-16}$. Compute $(1 \quad -1 \quad 0) \cdot \begin{pmatrix} 1 + \varepsilon \\ 1 \\ 1 \end{pmatrix} = (1 + \varepsilon) - 1 + 0 = \varepsilon$.

In this case we have a subtraction between two very close numbers.

$|\text{computer result}| \leq u \cdot (1 + |1 \cdot (1 + \varepsilon)| + |-1 \cdot 1| + |0 \cdot 1|) = (2 + \varepsilon) \cdot u$, which means that we have 10 correct digits.

The problem is in the fact that we have a very small result ($\varepsilon = O(10^{-6})$) and a very small error (of the same order of the result), which implies a large relative error.

An attentive reader might have noticed that it's very long to make this computation. Therefore, since we computed it on easy examples, we have to observe that we can't afford it for SVD or other complicated methods.

Wilkinson trick (from the 60's) comes to help us, to simplify this computation for more complicated problems.

Idea: see the computer result as the exact output of an algorithm run on a slightly perturbed input.

$$\begin{aligned}\tilde{y} &= \dots \\ &= ((a_1 b_1(1 + \delta_1) + a_2 b_2(1 + \delta_2))(1 + \delta_4) + a_3 b_3(1 + \delta_3))(1 + \delta_5) \quad (10.1.3) \\ &= a_1 \tilde{b}_1 + a_2 \tilde{b}_2 + a_3 \tilde{b}_3\end{aligned}$$

where

$$\begin{aligned}\tilde{b}_1 &= b_1(1 + \delta_1)(1 + \delta_4)(1 + \delta_5) = b_1 + 3ub_1 + o(u), \\ \tilde{b}_2 &= b_2(1 + \delta_2)(1 + \delta_4)(1 + \delta_5) = b_2 + 3ub_2 + o(u), \\ \tilde{b}_3 &= b_3(1 + \delta_2)(1 + \delta_5) = b_3 + 2ub_3 + o(u),\end{aligned}\quad (10.1.4)$$

$$\tilde{a}_i = a_i \quad i = 1, 2, 3.$$

And the relative error is $\frac{\|\tilde{b} - b\|}{\|b\|} \leq 3u + o(u)$.
Hence,

$$\frac{\|\tilde{y} - y\|}{\|y\|} \leq \kappa_{rel,b} \frac{\|\tilde{b} - b\|}{\|b\|} \leq k_{rel,b} \cdot 3u$$

This isn't bad, because it's within a factor 3 of the optimal error for a perfect algorithm.

Computing the conditioning is much easier than making all the calculations of δ s.

Definition 10.1.2 (Backward stability of an algorithm). *An algorithm to compute $\mathbf{y} = f(\mathbf{x})$ is called **backward stable** if the computed output $\tilde{\mathbf{y}}$ can be written as $\tilde{\mathbf{y}} = f(\tilde{\mathbf{x}})$, where $\tilde{\mathbf{x}} = \mathbf{x} + O(\mathbf{u} \|\mathbf{x}\|)$ (exact function, perturbed input).*

Observation 10.1.1. *In real-life usage, this $O()$ notation often hides polynomial factors in the dimension n . Although this may look an illicit simplification, we observe that these factors are much more harmless than the error that we could make otherwise.*

Theorem 10.1.1. *Backward stable algorithms are as accurate as theoretically possible (given the condition number of a problem), up to some factor that depends only on the dimension (e.g., n , $2n^2 + 18n$, ...).*

Proof.

$$\frac{\|\tilde{\mathbf{y}} - \mathbf{y}\|}{\|\mathbf{y}\|} \leq \kappa_{rel}(f, \mathbf{x}) \frac{\|\tilde{\mathbf{x}} - \mathbf{x}\|}{\|\mathbf{x}\|} = \kappa_{rel}(f, \mathbf{x})O(\mathbf{u}),$$

while the best attainable accuracy is $\kappa_{rel}(f, \mathbf{x})\mathbf{u}$. \square

We may ask ourselves if it's possible to perturb the input in order to get \tilde{y} for every possible algorithm and the answer is no. Let us see a counterexample, where $f(a, b) = a^T b$ (vector-vector product in the order that produces a matrix).

We observe that in general it's not true that the computed approximation of $f(a, b)$ has rank 1.

$$\begin{bmatrix} x_1 \\ x_2 \\ x_3 \end{bmatrix} \cdot \begin{bmatrix} y_1 & y_2 & y_3 \end{bmatrix} = \begin{bmatrix} x_1 y_1 & x_1 y_2 & x_1 y_3 \\ x_2 y_1 & x_2 y_2 & x_2 y_3 \\ x_3 y_1 & x_3 y_2 & x_3 y_3 \end{bmatrix} = M$$

While

$$\begin{bmatrix} x_1 \\ x_2 \\ x_3 \end{bmatrix} \odot \begin{bmatrix} y_1 & y_2 & y_3 \end{bmatrix} = \begin{bmatrix} x_1 \odot y_1 & x_1 \odot y_2 & x_1 \odot y_3 \\ x_2 \odot y_1 & x_2 \odot y_2 & x_2 \odot y_3 \\ x_3 \odot y_1 & x_3 \odot y_2 & x_3 \odot y_3 \end{bmatrix} = \widetilde{M}$$

And we are looking for \tilde{x} and \tilde{y} such that $\widetilde{M} = \tilde{x}^T \cdot \tilde{y}$

Example 10.1.2. Example (with exaggerated errors):

$$\begin{bmatrix} 1 \\ 2 \\ 3 \end{bmatrix} \cdot [4 \quad 5 \quad 6] = \begin{bmatrix} 4 & 5 & 6 \\ 8 & 10 & 12 \\ 12 & 15 & 18 \end{bmatrix}$$

While

$$\begin{bmatrix} 1 \\ 2 \\ 3 \end{bmatrix} \odot [4 \quad 5 \quad 6] = \begin{bmatrix} 4.01 & 4.99 & 6.01 \\ 7.99 & 10.01 & 12.02 \\ 11.98 & 15.02 & 17.97 \end{bmatrix}$$

and this second matrix doesn't have rank 1.

10.2 Backward stability of QR factorization

💡 Do you recall?

A generic step of the computation of the QR factorization of a matrix has the following shape:

$$\begin{pmatrix} & I \\ & H \ u_k \end{pmatrix} \begin{pmatrix} * & * & * & * & * \\ 0 & * & * & * & * \\ 0 & 0 & * & * & * \\ 0 & 0 & * & * & * \\ 0 & 0 & * & * & * \end{pmatrix} = \begin{pmatrix} * & * & * & * & * \\ 0 & * & * & * & * \\ 0 & 0 & * & * & * \\ 0 & 0 & 0 & * & * \\ 0 & 0 & 0 & * & * \end{pmatrix}$$

Each step of the QR factorization is backward stable.

Let us compute the backward stability of QR factorization:

With some tedious computations like the ones above, one can show that *one* step of the QR factorization is backward stable, i.e., $\tilde{R}_{k-1} = R_{k-1} + \Delta_{R_{k-1}}$, where $\|\Delta_{R_{k-1}}\| \leq O(u) \|R_{k-1}\|$ and this allows us to write the following

$$\frac{\|\tilde{R}_{k-1} - R_{k-1}\|}{\|R_{k-1}\|} \leq O(u)$$

At a generic step k we have that

$$R_k = Q_k \cdot R_{k-1}$$

\Updownarrow

$$\tilde{R}_k = Q_k(R_{k-1} + \Delta_{R_{k-1}})$$

The idea is doing this procedure one step after the other, starting from step 1, where $R_0 = A$.

$$\tilde{R}_1 = Q_1 \tilde{R}_0 = Q_1 \cdot (R_0 + \Delta_{R_0}) = Q_1 \cdot (A + \Delta_{R_0})$$

$$\tilde{R}_2 = Q_2 \tilde{R}_1 = Q_2(R_1 + \Delta_{R_1}) = Q_2 Q_1(A + \Delta_{R_0}) + Q_2 \Delta_{R_1} = Q_2 Q_1(A + \Delta_{R_0} + Q_1^T \Delta_{R_1})$$

We go on combining errors like this and we see that all the quantities $(A + \Delta_{R_j})$ are perturbations of the original matrix A .

In the end we may observe that the final computed R_n is the exact result obtained from A plus n perturbations.

Observation 10.2.1. *We should notice that the norm of each perturbation is small with respect to the norm of A .*

In formal terms, $\|\Delta_{R_0}\| \leq u \cdot \|R_0\| = u \cdot \|A\|$ and $\|Q_1^T \Delta_{R_1}\| \stackrel{(1)}{=} \|\Delta_{R_1}\| \leq u \cdot \|R_1\| \stackrel{(2)}{=} u \cdot \|A\|$.

Where $\stackrel{(1)}{=}$ and $\stackrel{(2)}{=}$ hold because Q_i are orthogonal.

★ Mantra

Orthogonal transformations are the key for stability.

10.2.1 Stability of algorithms for least-squares problems

Let us see how various algorithms to solve LS problems (implemented in Matlab) behave in relation to backward stability.

Least squares problem via QR

STEP 1: Computing a thin QR (`qr(A, 0);`) → backward stable;

STEP 2: $(Q_1' * b;)$ → backward stable;

STEP 3: $(R_1 \ c;)$ → backward stable;

Least squares problem via SVD

Scrivere meglio

STEP 1: Computing a SVD ($\text{svd}(A, 0;)$) → backward stable;

STEP 2: $(U' * b;)$ → backward stable;

STEP 3: $(c ./ \text{diag}(S);)$ → backward stable;

STEP 4: $(V * d;)$ → backward stable;

Least squares problem via Normal Equations

STEP 1: $C = A' * A;$

STEP 2: $d = A' * b;$

STEP 3: $x = C^{-1} d;$

We can also prove that some algorithms aren't backward stable, like normal equations. The issue here is that the same input is needed in more than one operation, so it should satisfy more than one equation and this may lead to a solution set which is empty. Also, in the last equation we solve a linear system with matrix C , and this gives an error of the order of $\kappa(A)^2$, never of $\kappa(A)$: $\kappa(C) = \kappa(A^T A) = \kappa(A)^2$.

Let us take the example of two lectures ago: ??.

Example 10.2.1. Let $A \in \mathcal{M}(4, 3, \mathbb{R})$ s.t.

$$A = \begin{pmatrix} 1 & 1 & 2 \\ 1 & 2 & 3 \\ 3 & 1 & 4 \\ 1 & 2 & 3 + 10^{-8} \end{pmatrix}$$

We may observe that A is at distance 10^{-8} from a matrix without full column rank, hence $\kappa \approx 10^8$.

What is the condition number of solving this least squares problem? It's about 10^8 , since in that problem we generated $b = A \cdot (3 \ 4 \ 5)$, so b lies in the image of A ($\text{Im}(A)$). The geometric idea in Figure 10.2.

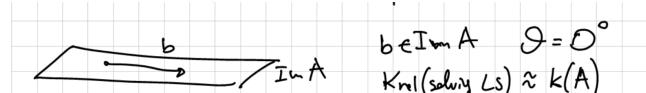


FIGURE 10.2: b lies (at least almost, because of numerical error) on the plane of $\text{Im}(A)$. In this case $\kappa_{\text{rel}}(\text{solving LS}) \approx \kappa(A)$.

A brief recap may be found in Table 10.1.

	Normal eqns	QR	SVD
$m \approx n$	$\frac{4}{3}n^3$	$\frac{4}{3}n^3$	$\approx 13n^3$
$m \gg n$	mn^2	$2mn^2$	$2mn^2$
Unstable when $cond \approx \kappa(A)$	Backward stable	Backward stable	Backward stable; reveals info on sensitivity, allows regularization

TABLE 10.1: Brief recap of the complexities of the algorithms we studied for solving the least squares problem. The last row takes into account the stability.

10.2.2 A posteriori checks

Let's assume that Matlab gives us a solution of a problem. We want to be able to check how good this result is.

Example 10.2.2. Suppose we have solved a linear system...

```
>> A = randn(4, 4); b = randn(4, 1);
>> x = A \ b;
>> A*x - b
ans =
    0
   -1.3878e-17
    0
   2.2204e-16
```

Definition 10.2.1 (Residual). Let $A \in \mathcal{M}(m, \mathbb{R})$ and $\in \mathbb{R}^m$, and x be the solution of $Ax = b$. For a given \tilde{x} we define **residual** the following $\mathbf{r} = A\tilde{x} - b$.

assume that $\|\mathbf{r}\|$ is small; does this mean that \tilde{x} is close to the exact solution x ?

Theorem 10.2.1. Let $A \in \mathbb{R}^{m \times m}, \in \mathbb{R}^m$, and x be the solution of $Ax = b$.

For a given \tilde{x} , the relative error of \tilde{x} is bounded by the condition of matrix A times the ratio between the norm of the residual and the norm of b .

$$\frac{\|x - \tilde{x}\|}{\|x\|} \leq \kappa(A) \frac{\|\mathbf{r}\|}{\|b\|}.$$

This theorem tells us that \tilde{x} is “close to the solution” apart from a factor which is the conditioning of matrix A .

Proof. Follows from the perturbation results for linear systems. The idea is that \tilde{x} is the exact solution of the perturbed system

$$A\tilde{x} \stackrel{*}{=} b + \mathbf{r} = \tilde{b}$$

* Where = follows from the definition of r and $\frac{\|\tilde{b}-b\|}{\|b\|} = \frac{\|r\|}{\|b\|}$.
A relative perturbation of size $\frac{r}{b}$ is amplified by $\kappa(A)$. \square

It's important to notice that also computing $A \odot x \ominus b$ is an approximated operation. We choose to simplify things and ignore this error.

10.3 A posteriori check for Least Squares Problems

Can we make an analogous check for the Least Squares Problem?

Take a LSP $\min \|Ax - b\|$, with A tall thin, with full column rank.

The problem is that $\|Ax - b\|$ isn't small at all, indeed it could be as large as b , as you can see from Figure 10.3.

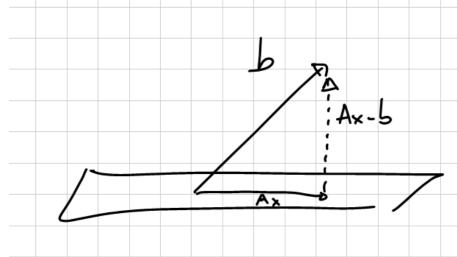


FIGURE 10.3: A can be as large as b if b is perfectly orthogonal.

Observation 10.3.1. If you solve LSP via QR $\|Ax - b\| = \left\| \begin{pmatrix} R_1 x - Q_1^T b \\ -Q_2^T b \end{pmatrix} \right\|$.

We said that the entries in the second block are fixed irrespective of x , but we could make the entries in the first block zero, by choosing $x = R_1^{-1} Q^T b$. This information let us infer something about the values of the vectors in Figure 10.3, in particular the minimum of the value that we can get is $\|Q_2^T b\| = \|Ax - b\|$.

With some algebra we may also check that $\|Q_1^T b\| = \|Ax\|$.

Since this is a minimum problem, we know that the gradient of the function is small near the optimum value: $\min \|Ax - b\|^2 = \min x^T A^T Ax - 2b^T Ax + b^T b$.

$$\nabla_{\tilde{x}} f = 2(A^T Ax - A^T b) \rightarrow 0$$

Theorem 10.3.1. $\frac{\|\tilde{x} - x\|}{\|x\|} \leq (\kappa(A))^2 \cdot \frac{\|A^T Ax - A^T b\|}{\|A^T b\|}$. Although we might have wanted to have the condition number of the problem, instead of the condition number of A and this could lead to underestimating the error.

Another idea could be using as error the first entry of the vector obtained via QR (namely $R_1^{-1} x - Q_1^T b$), by imposing $R_1 x = Q_1^T b$

We may observe that this is a truly backward stable measure:
given $r = \|R_1^T \tilde{x} - Q_1^T b\|$, there exists \tilde{b} with $\|\tilde{b} - b\| = \|r\|$ such that \tilde{x} is
the exact solution of $\min \|Ax - \tilde{b}\|$.

We have proved the following

$$\text{Fact 10.3.2. } \frac{\|\tilde{x} - x\|}{\|x\|} \leq \kappa_{rel,LS} \cdot \frac{\|\tilde{b} - b\|}{\|b\|} = \kappa_{rel,LS} \cdot \frac{\|r\|}{\|b\|}.$$

Theorem 10.3.3. Let $A = Q_1 R_1$ be a thin QR factorization. Let $\mathbf{r}_1 = Q_1^T(A\tilde{x} - b)$. Then, \tilde{x} is the exact solution of the LS problem

$$\min \|A\mathbf{x} - (\mathbf{b} + Q_1 \mathbf{r}_1)\|,$$

so the backward error of \tilde{x} is $\|Q_1 \mathbf{r}_1\| = \|\mathbf{r}_1\|$.

Proof. Idea: replay the solution of a LS problem with QR factorization, and
use $Q_1^T T Q_1 = I$. You will get in the first block $R_1 x = Q_1^T b + \mathbf{r}_1$, i.e.,
 $Q_1^T(Ax - b) = \mathbf{r}_1$, which is verified by \tilde{x} . \square

Lecture 11

21st of November 2018 — F. Poloni

In this lecture we address the problem of solving linear systems exactly.

Someone could observe that this subject has already been studied in the numerical linear algebra course, but we are interested in computing the solution to this problem quickly when the dimensions are large and the matrix A is sparse.

Since the complexity of Gauss method is cubic, this algorithm is unfeasible for large inputs.

Let us see some real life examples, where the matrices are large and sparse.

LOCAL FUNCTION ON GRAPHS: A **local function on graphs** is a function that depend on few nearby vertices. This kind of functions lead to a sparse adjacency matrix A , as can be observed in Figure 11.1;

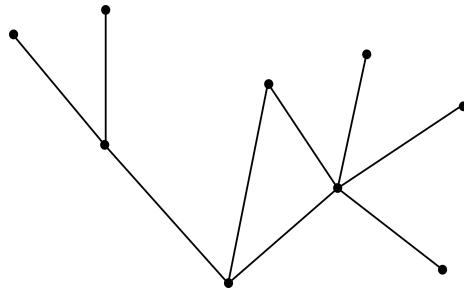


FIGURE 11.1: A local function on graph.

IMAGES: Take an $m \times m$ image and blur it (each pixel is obtained as the average of its neighbours). $T : \mathcal{M}(m, \mathbb{R}) \rightarrow \mathcal{M}(m, \mathbb{R})$ such that $T(A)_{ij} = \frac{1}{9} \cdot (A_{i-1,j} + A_{i-1,j-1} + A_{i-1,j+1} + A_{ij} + A_{ij-1} + A_{ij+1} + A_{i+1,j} + A_{i+1,j-1} + A_{i+1,j+1})$.

T may be written as a matrix that maps all the m images to a set of m blurred images and has the following shape $T \in \mathcal{M}(m^2, \mathbb{R})$ such that the (i, j) -th row of T has exactly 9 entries with value $\frac{1}{9}$ and all the others are 0. The non zero entries correspond to $A_{i-1j}, \dots, A_{i+1j+1}$;

KKT SYSTEMS constrained optimization;

ENGINEERING PROBLEM: To check stability of a bridge, it gets split into small cells. It can be proven that the stress on each of these cells corresponds to the force applied by the neighbours. In the end, this local phenomenon may be represented by a sparse matrix.

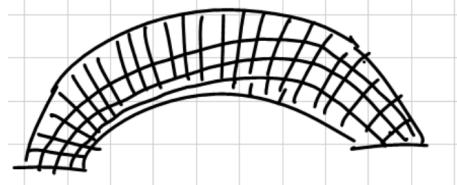


FIGURE 11.2: Graphic idea of a bridge partitioned into small blocks

11.1 Gaussian elimination and LU factorization

Gaussian elimination can be seen as a factorization: $A = LU$. The intuition is to proceed iteratively, multiplying each time for a new matrix, just like QR factorization.

Since the idea of Gauss elimination is to add multiples of row 1 to all the rows from 2 to n to kill off $A_{2:end,1}$ we have that:

STEP 1:

$$\begin{pmatrix} 1 & & & & \\ * & 1 & & & \\ * & & 1 & & \\ * & & & 1 & \\ * & & & & 1 \end{pmatrix} \begin{pmatrix} \textcircled{*} & * & * & * & * \\ * & * & * & * & * \\ * & * & * & * & * \\ * & * & * & * & * \\ * & * & * & * & * \end{pmatrix} = \begin{pmatrix} \textcircled{*} & * & * & * & * \\ 0 & * & * & * & * \\ 0 & * & * & * & * \\ 0 & * & * & * & * \\ 0 & * & * & * & * \end{pmatrix},$$

Where the $\textcircled{*}$ is called **pivot**.

$$L_1 A = A_1.$$

$$(L_1)_{k1} = -\frac{A_{k1}}{A_{11}}, \quad k = 2, 3, \dots, m.$$

STEP 2: we multiply for a matrix that has an “identity frame” and inside does the same L_1 was doing before.

$$\begin{pmatrix} 1 & & & & \\ & 1 & & & \\ * & 1 & & & \\ * & & 1 & & \\ * & & & 1 & \end{pmatrix} \begin{pmatrix} * & \odot & * & * & * \\ 0 & * & * & * & * \\ 0 & * & * & * & * \\ 0 & * & * & * & * \\ 0 & * & * & * & * \end{pmatrix} = \begin{pmatrix} * & \odot & * & * & * \\ 0 & * & * & * & * \\ 0 & 0 & * & * & * \\ 0 & 0 & * & * & * \\ 0 & 0 & * & * & * \end{pmatrix}$$

$$L_2 A_1 = A_2,$$

$$(L_2)_{k2} = \frac{(A_1)_{k2}}{(A_1)_{22}}, \quad k = 3, \dots, m.$$

STEP 3: we go on and

$$\begin{pmatrix} 1 & & & & \\ & 1 & & & \\ & & 1 & & \\ & * & 1 & & \\ * & & & 1 & \end{pmatrix} \begin{pmatrix} * & * & * & * & * \\ 0 & * & * & * & * \\ 0 & 0 & \odot & * & * \\ 0 & 0 & * & * & * \\ 0 & 0 & * & * & * \end{pmatrix} = \begin{pmatrix} * & * & * & * & * \\ 0 & * & * & * & * \\ 0 & 0 & \odot & * & * \\ 0 & 0 & 0 & * & * \\ 0 & 0 & 0 & * & * \end{pmatrix}$$

$$L_3 A_2 = A_3,$$

$$(L_3)_{k3} = \frac{(A_2)_{k3}}{(A_2)_{33}}, \quad k = 4, \dots, m.$$

STEP 4: one more operation

$$\begin{pmatrix} 1 & & & & \\ & 1 & & & \\ & & 1 & & \\ & & & 1 & \\ * & & & & 1 \end{pmatrix} \begin{pmatrix} * & * & * & * & * \\ 0 & * & * & * & * \\ 0 & 0 & * & * & * \\ 0 & 0 & 0 & \odot & * \\ 0 & 0 & 0 & * & * \end{pmatrix} = \begin{pmatrix} * & * & * & * & * \\ 0 & * & * & * & * \\ 0 & 0 & * & * & * \\ 0 & 0 & 0 & \odot & * \\ 0 & 0 & 0 & 0 & * \end{pmatrix}$$

$$L_4 A_3 = A_4,$$

$$(L_4)_{k4} = \frac{(A_3)_{k4}}{(A_3)_{44}}, \quad k = 5, \dots, m.$$

In the generic case we have $L_{m-1} L_{m-2} \dots L_1 A = U$, where U is upper triangular, or $A = \underbrace{L_1^{-1} L_2^{-1} \dots L_{m-1}^{-1}}_{=L} U$, with U upper triangular and L lower triangular.

Theorem 11.1.1. Let $A \in \mathcal{M}(m, \mathbb{R})$ such that we do not encounter zero pivots in the algorithm. A admits a factorization $A = LU$, where L is lower triangular with ones on its diagonal, and U is upper triangular.

Observation 11.1.1 (Stroke of luck). *The product of the L_i^{-1} 's (denoted L) can be computed for free, since the following holds:*

$$\left[\begin{array}{cc} 1 & \\ -a_2 & 1 \\ -a_3 & 1 \\ -a_4 & 1 \\ -a_5 & 1 \end{array} \right]^{-1} \cdot \left[\begin{array}{cc} 1 & \\ & 1 \\ -b_3 & 1 \\ -b_4 & 1 \\ -b_5 & 1 \end{array} \right]^{-1} \cdot \left[\begin{array}{cc} 1 & \\ & 1 \\ & -c_4 & 1 \\ & -c_5 & 1 \end{array} \right]^{-1} \cdot \left[\begin{array}{cc} 1 & \\ & 1 \\ & & 1 \\ & & -d_5 & 1 \end{array} \right]^{-1} = \left[\begin{array}{ccccc} 1 & & & & \\ a_2 & 1 & & & \\ a_3 & b_3 & 1 & & \\ a_4 & b_4 & c_4 & 1 & \\ a_5 & b_5 & c_5 & d_5 & 1 \end{array} \right]$$

ALGORITHM 11.1.1 LU factorization, Matlab implementation.

```

1  function [L, U] = lu_factorization(A)
2      m = size(A, 1);
3      L = eye(m);
4      U = A;
5      for k = 1 : m - 1
6          % compute "multipliers"
7          L(k+1:end, k) = U(k+1:end, k) / U(k, k);
8          % update U
9          U(k+1:end, k) = 0;
10         U(k+1:end, k+1:end) = U(k+1:end, k+1:end) ...
11             - L(k+1:end, k) * U(k, k+1:end);
12     end

```

The idea behind the implementation of Algorithm 11.1.1 is shown in Figure 11.3.

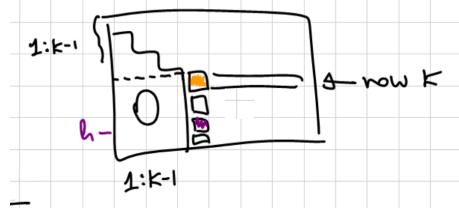


FIGURE 11.3: Assuming that we are at step k , we have that the h -th multiplier is expressed as $\frac{A_{hk}}{A_{kk}}$ and this multiplier goes to the right position in L .

Observation 11.1.2. *The computational complexity of this algorithm is concentrated at lines 10, 11 and the cost of this operation is $O((m-1)^2 + (m-2)^2 + \dots + 2^2 + 1)$. So we have that for a dense matrix the computational complexity is $\frac{2}{3}m^3 + O(m^2)$, in other words half as much as QR factorization.*



Something on Matlab ...

Implementation of \ in Matlab: We consider important to remark how the operator \backslash is implemented in Matlab. It works using the LU factorization, making some checks and changing the algorithm as follows:

- upper/lower triangular systems: back-substitution ($O(n^2)$);
- non-triangular linear systems: LU with partial pivoting (then throw away the factors);
- symmetric and/or sparse systems: uses appropriate LU variants (will see in the following).
- non-square matrices: solves the system “in the least squares sense $\min_x \|Ax - b\|_2$ ”.

Why isn't there any check on the orthogonality of the matrix? Because in that case $A^{-1} = A^T$, and hence it is easy to solve the system. Well, in order to find out that a matrix is orthogonal we need to compute $A^T A$, which is too costly.

Obs: LU without pivoting isn't stable, as shown in Section 11.1.1, so Matlab uses pivoting.

missing

11.1.1 Stability of LU

A downside of this approach is that it's not numerically stable. The intuition is that the condition is bad whenever the matrix A has a very small pivot.

Let us see an example:

Example 11.1.1.

$$A = \begin{bmatrix} 10^{-30} & 1 \\ 1 & 1 \end{bmatrix} = \begin{bmatrix} 1 & 0 \\ 10^{30} & 1 \end{bmatrix} \begin{bmatrix} 10^{-30} & 1 \\ 0 & 1 - 10^{30} \end{bmatrix}.$$

In this case the LU factorization produces L, U with norm much larger than $\|A\|$.

Luckily, it's easy to circumvent this issue multiplying L_i s by some permutation matrices (which swap rows in order to keep “large” pivots), as follows

$$L_{m-1} P_{m-1} \dots L_2 P_2 L_1 P_1 A = U.$$

Observation 11.1.3. Thanks to another “stroke of luck” we can reorder those factors:

$$L_{m-1} P_{m-1} \dots L_2 P_2 L_1 P_1 = \widehat{L}_{m-1} \widehat{L}_{m-2} \dots \widehat{L}_1 P_{m-1} P_{m-2} \dots P_1$$

where \widehat{L}_i have the same structure as the L_i .

We can now introduce the following

Theorem 11.1.2. Let $A \in \mathcal{M}(m, \mathbb{R})$. A admits a factorization $A = PLU$, where P is a permutation matrix, L is lower triangular with ones on its diagonal, and U is upper triangular.

What about the stability of this improvement to LU factorization, called **LU with partial pivoting**?

It's not stable at all, in the worst case it may happen that $\|U\|/\|A\|$ may grow as $\approx 2^m$, although matrices for which this happens are very rare.

11.1.2 Gaussian elimination on sparse matrices

Given a sparse matrix

$$A = \begin{pmatrix} * & * & & * & * & * \\ * & * & * & * & * & * \\ * & * & * & * & * & * \\ * & * & * & * & * & * \\ * & * & * & * & * & * \\ * & * & * & * & * & * \\ * & * & * & * & * & * \\ * & * & * & * & * & * \end{pmatrix}$$

Gaussian elimination causes some fill-in, due to the sum of a multiple of the first row, which has non zero entries in different positions:

$$\begin{pmatrix} * & * & & * & * & * \\ * & * & * & * & * & * \\ * & * & * & * & * & * \\ * & * & * & * & * & * \\ * & * & * & * & * & * \\ * & * & * & * & * & * \\ * & * & * & * & * & * \\ * & * & * & * & * & * \end{pmatrix} \rightarrow \begin{pmatrix} * & * & & * & * & * \\ * & * & * & * & * & * \\ 0 & * & * & * & * & * \\ 0 & * & * & * & * & * \\ * & * & * & * & * & * \\ 0 & * & * & * & * & * \\ 0 & * & * & * & * & * \\ 0 & * & * & * & * & * \end{pmatrix}$$

We may observe that the computational complexity of sparse LU is linear in the number of non zero entries of the final matrix, obtained by the algorithm, which is possibly much larger than the number of non zeros in A .

How to circumvent this problem? At each step we may use as pivot row the most sparse one. This computation may be done in a more sofisticate way, considering the “relative” position of non zeros between couples of rows.

Because of this trade-off the choice is made in relation to the needs of the implementation. We won't study any algorithm that deals with sparse matrices, since they are very complicated and make use of heuristics.

There are some lucky cases in which the fill-in is almost none, for example a matrix that only has 5 diagonals which entries are different from 0 (called **tridiagonal**). In this particular case L is tridiagonal and lower triangular and U is tridiagonal and upper triangular, as shown below.

$$A = \begin{pmatrix} * & * & * & 0 & 0 & 0 & \cdots & 0 \\ * & * & * & * & 0 & 0 & \cdots & 0 \\ * & * & * & * & * & 0 & \cdots & 0 \\ 0 & * & * & * & * & * & & \vdots \\ \vdots & \ddots & \ddots & \ddots & \ddots & \ddots & \ddots & 0 \\ 0 & 0 & 0 & \cdots & * & * & * & * \\ 0 & 0 & 0 & \cdots & * & * & * & * \\ 0 & 0 & 0 & \cdots & 0 & * & * & * \end{pmatrix}$$

$$L = \begin{pmatrix} * & & & & & & \\ * & * & & & & & \\ * & * & * & & & & \\ & \ddots & \ddots & \ddots & & & \\ & & * & * & * & & \\ & & & * & * & * & \\ & & & & * & * & * \end{pmatrix} U = \begin{pmatrix} * & * & * & & & & \\ * & * & * & & & & \\ * & * & * & & & & \\ & \ddots & \ddots & \ddots & & & \\ & & * & * & * & & \\ & & & * & * & * & \\ & & & & * & * & * \end{pmatrix}$$

Observation 11.1.4. *We should remark that if we are interested in high-performance computing we need to pay attention to the blocking, because we go from vector-vector operation to matrix-matrix operation and some of these operations may be performed more efficiently. Parallel/multithreaded implementations are available by means of parallel libraries for Matlab.*

Lecture 12

23rd of November 2018 —
F. Poloni

12.1 Gaussian elimination on symmetric matrices

💡 Do you recall?

In Gaussian elimination we had A and we multiplied it by L_1 in order to get a big chunk of 0s in the first column

$$\begin{pmatrix} 1 & & & & \\ * & 1 & & & \\ * & & 1 & & \\ * & & & 1 & \\ * & & & & 1 \end{pmatrix} \begin{pmatrix} (\otimes) & * & * & * & * \\ * & * & * & * & * \\ * & * & * & * & * \\ * & * & * & * & * \\ * & * & * & * & * \end{pmatrix} = \begin{pmatrix} (\otimes) & * & * & * & * \\ 0 & * & * & * & * \\ 0 & * & * & * & * \\ 0 & * & * & * & * \\ 0 & * & * & * & * \end{pmatrix},$$

Let us consider an upgrade of Gaussian elimination in the case of $A \in S(m, \mathbb{R})$.

Let us see what happens if we multiply $L_1 A$ on the right by the transpose of L_1 :

STEP 1:

$$\begin{pmatrix} 1 & & & & \\ * & 1 & & & \\ * & & 1 & & \\ * & & & 1 & \\ * & & & & 1 \end{pmatrix} \begin{pmatrix} * & * & * & * & * \\ * & * & * & * & * \\ * & * & * & * & * \\ * & * & * & * & * \\ * & * & * & * & * \end{pmatrix} \begin{pmatrix} 1 & * & * & * & * \\ & 1 & & & \\ & & 1 & & \\ & & & 1 & \\ & & & & 1 \end{pmatrix} = \begin{pmatrix} * & 0 & 0 & 0 & 0 \\ 0 & * & * & * & * \\ 0 & * & * & * & * \\ 0 & * & * & * & * \\ 0 & * & * & * & * \end{pmatrix}$$

STEP 2:

$$\begin{pmatrix} 1 & & & & \\ & 1 & & & \\ * & & 1 & & \\ * & & & 1 & \\ * & & & & 1 \end{pmatrix} \begin{pmatrix} * & 0 & 0 & 0 & 0 \\ 0 & * & * & * & * \\ 0 & * & * & * & * \\ 0 & * & * & * & * \\ 0 & * & * & * & * \end{pmatrix} \begin{pmatrix} 1 & & & & \\ & 1 & * & * & \\ & & 1 & & \\ & & & 1 & \\ & & & & 1 \end{pmatrix} = \begin{pmatrix} * & 0 & 0 & 0 & 0 \\ 0 & * & 0 & 0 & 0 \\ 0 & 0 & * & * & * \\ 0 & 0 & * & * & * \\ 0 & 0 & * & * & * \end{pmatrix}$$

STEP m :

$$L_{m-1} L_{m-2} \dots L_1 A L_1^T \dots L_{m-2}^T L_{m-1}^T = D,$$

where D is diagonal, or

$$A = L_1 L_2 \dots L_{m-1} D L_{m-1}^T \dots L_2^T L_1^T = LDL^T.$$

Observation 12.1.1 (Stroke of luck). Notice that the stroke of luck of Observation 11.1.1 holds in this case too, hence we pay nothing to compute matrix L .

$$\begin{bmatrix} 1 & & & & \\ -a_2 & 1 & & & \\ -a_3 & & 1 & & \\ -a_4 & & & 1 & \\ -a_5 & & & & 1 \end{bmatrix}^{-1} \cdot \begin{bmatrix} 1 & & & & \\ -\frac{1}{b_3} & 1 & & & \\ -\frac{b_4}{b_3} & & 1 & & \\ -\frac{b_5}{b_3} & & & 1 & \\ & & & & 1 \end{bmatrix}^{-1} \cdot \begin{bmatrix} 1 & & & & \\ & 1 & & & \\ & & 1 & & \\ & & -c_4 & 1 & \\ & & -c_5 & & 1 \end{bmatrix}^{-1} \cdot \begin{bmatrix} 1 & & & & \\ & 1 & & & \\ & & 1 & & \\ & & & 1 & \\ & & & & 1 \end{bmatrix}^{-1} = \begin{bmatrix} 1 & & & & \\ a_2 & 1 & & & \\ a_3 & b_3 & 1 & & \\ a_4 & b_4 & c_4 & 1 & \\ a_5 & b_5 & c_5 & d_5 & 1 \end{bmatrix}$$

Theorem 12.1.1 (Symmetric Gaussian elimination). Let $A \in S(m, \mathbb{R})$ such that during Gaussian elimination we don't encounter any 0 pivot. A admits a factorization $A = LDL^T$, where L is lower triangular with ones on its diagonal, and D is diagonal.

A Matlab implementation of symmetric Gaussian elimination is shown in Algorithm 12.1.1.

ALGORITHM 12.1.1 Symmetric Gaussian factorization, Matlab implementation.

```

1  function [L, D] = ldl_factorization(A)
2      m = size(A, 1);
3      L = eye(m); D = zeros(m);
4      for k = 1:m-1
5          D(k, k) = A(k, k);
6          L(k+1:end, k) = A(k+1:end, k) / A(k, k);
7          A(k+1:end, k+1:end) = A(k+1:end, k+1:end) ...
8              - L(k+1:end, k) * A(k, k+1:end);
9      end
10     D(m, m) = A(m, m);

```

It is possible to make an optimization of this algorithm: since A is supposed to be symmetric, we only need to update the lower triangular part of A , since the rest is mirrored by symmetry, hence the computational complexity is half that of Gaussian elimination.

This algorithm is not backward stable, exactly like the one on non symmetric matrices. Pivoting may be performed in order to improve stability. It comes without saying that the row swap should be done consistently on the columns to preserve symmetry.

Of course there are some matrices (like the ones with all 0s on the diagonal) that cannot be “pivoted”. There are workarounds, though. As an example, Matlab’s $[L, D, P] = \text{ldl}(A)$ produces matrices such that $P^T AP = LDL^T$, where D may have 2×2 diagonal blocks.

💡 Do you recall?

We recall the characterization of **positive definite matrix** $A \in M(m, \mathbb{R})$: all its eigenvalues are strictly positive. In other words, A is positive definite if $\forall z \neq 0 \in \mathbb{R}^m z^T Az > 0$.

Lemma 12.1.2. *In the context of positive definite matrices the following holds:*

1. *Let A be a symmetric matrix. A is positive definite if and only if MAM^T is so, for some invertible $M \in M(m, \mathbb{R})$. Formally, $\forall A \in S(m, \mathbb{R})$ s.t. $A\succ 0 \Leftrightarrow \exists M \in GL(m, \mathbb{R})$ s.t. $MAM^T\succ 0$;*
2. *Let A a symmetric positive definite matrix such that $A = \begin{bmatrix} A_{11} & A_{12} \\ A_{21} & A_{22} \end{bmatrix}$, then A_{11} and A_{22} are, too. Formally, $\forall A \in S(m, \mathbb{R})$ s.t. $A\succ 0$ and $A = \begin{bmatrix} A_{11} & A_{12} \\ A_{21} & A_{22} \end{bmatrix} \Rightarrow A_{11}\succ 0$ and $A_{22}\succ 0$.*

Proof.

1.

$$\Rightarrow) A \in S(m, \mathbb{R}) \text{ and } A\succ 0 \implies MAM^T \in S(m, \mathbb{R}) \text{ and } MAM^T\succ 0.$$

Take $z \in \mathbb{R}^m$, $z \neq 0$ $z^T MAM^T z = y^T Ay > 0$, where we performed a variable change $y = M^T z$. Notice that $y \neq 0$ because M is invertible (and $\ker(M) = \{0\}$). The symmetry of the matrix MAM^T follows from $(MAM^T)^T = M^T A^T M^T = MAM^T$;

$$\Leftarrow) MAM^T \in S(m, \mathbb{R}) \text{ and } MAM^T\succ 0 \implies A \in S(m, \mathbb{R}) \text{ and } A\succ 0.$$

This proof follows from the previous arrow, where the substitution is $z = M^{-1}y$.

$$2. A = \begin{bmatrix} A_{11} & A_{12} \\ A_{21} & A_{22} \end{bmatrix} \text{ positive definite} \implies A_{11} \text{ and } A_{22} \text{ are positive definite too.}$$

Since A is positive definite, its scalar product is greater than zero with all the vectors in \mathbb{R}^m .

A_{11}) Let us take $\mathbf{z} = \begin{bmatrix} \mathbf{z}_1 \\ 0 \end{bmatrix}$.

$$\begin{bmatrix} \mathbf{z}_1^T & 0 \end{bmatrix} \cdot \begin{bmatrix} A_{11} & A_{12} \\ A_{21} & A_{22} \end{bmatrix} \cdot \begin{bmatrix} \mathbf{z}_1 \\ 0 \end{bmatrix} = \mathbf{z}_1^T A_{11} \mathbf{z}_1 > 0, \forall \mathbf{z}_1 \in \mathbb{R}^{sizeof A_{11}}$$

A_{22}) Let us take $\mathbf{z} = \begin{bmatrix} 0 \\ \mathbf{z}_2 \end{bmatrix}$.

$$\begin{bmatrix} 0 & \mathbf{z}_2^T \end{bmatrix} \cdot \begin{bmatrix} A_{11} & A_{12} \\ A_{21} & A_{22} \end{bmatrix} \cdot \begin{bmatrix} 0 \\ \mathbf{z}_2 \end{bmatrix} = \mathbf{z}_2^T A_{22} \mathbf{z}_2 > 0, \forall \mathbf{z}_2 \in \mathbb{R}^{sizeof A_{22}}$$

□

Corollary 12.1.3. *Let $A \in M(n, \mathbb{R})$ such that A is positive definite. When computing the LDL^T factorization of A , at each step we have $D_{kk} > 0$, hence we need no pivoting technique.*

Proof. From the first point of Lemma 12.1.2 we have that, since $A \succ 0$, $L_1 A L_1^T$ is positive definite. Thanks to the second point of the same lemma we have

$$L_1 A L_1^T = \begin{pmatrix} * & 0 & 0 & 0 & 0 \\ 0 & * & * & * & * \\ 0 & * & * & * & * \\ 0 & * & * & * & * \\ 0 & * & * & * & * \end{pmatrix} \text{ and so the first and the second diagonal blocks}$$

are positive definite ($D_{11} > 0$ and $D_{22} \succ 0$).

Notice that, since A is positive definite $A_{11} \succ 0$, but it's a scalar, hence $A_{11} > 0$ and this implies no breakdown case. □

We may introduce another kind of factorization.

12.2 Cholesky factorization

The key idea is to write the diagonal matrix of the Gaussian elimination D as product of $D^{1/2}$ times itself:

$$D = \begin{pmatrix} d_{11} & & & & \\ & d_{22} & & & \\ & & \ddots & & \\ & & & d_{mm} & \end{pmatrix} = \begin{pmatrix} \sqrt{d_{11}} & & & & \\ & \sqrt{d_{22}} & & & \\ & & \ddots & & \\ & & & \sqrt{d_{mm}} & \end{pmatrix} \cdot \begin{pmatrix} \sqrt{d_{11}} & & & & \\ & \sqrt{d_{22}} & & & \\ & & \ddots & & \\ & & & \sqrt{d_{mm}} & \end{pmatrix}$$

The LDL factorization may be rewritten as follows

$$A = LDL^T = LD^{1/2}(D^{1/2}L^T) = CC^T,$$

where $D^{1/2} = \text{diag}(D_{11}^{1/2}, D_{22}^{1/2}, \dots, D_{mm}^{1/2})$, and C is lower triangular (but not anymore with ones on the diagonal).

In Matlab the Cholesky factorization of a positive definite matrix is performed by the function `chol(A)`; and returns C^T .

Observation 12.2.1. *We will not discuss stability further, but Cholesky is always backward stable even without pivoting ($\|C\| = \|A\|^{1/2}$).*

Observation 12.2.2. *In a sparse matrix, we can choose the (symmetric) permutation with the only goal of reducing fill-in. The same considerations about LU factorization hold in this case too.*

12.3 Krylov subspace methods

In this part we will discuss some different techniques to solved linear systems, inspired from optimization algorithms.

Example 12.3.1. Let us consider the optimization problem $\min \frac{1}{2}x^T Ax + b^T x$, where $A \succ 0$. We know that the solution to this problem is $x = A^{-1}b$. Assume we solve this problem via a gradient descent-type method, starting from $x_0 = 0$.

STEP 1: $x_0 = 0$, $\nabla f(x_0) = -b$;

STEP 2: $x_1 = \text{some multiple of } b \in \text{Span}(b)$

$$f(x_1) = Ax_1 - b$$

$$\nabla f(x_1) = \text{some multpiple of } Ab + \text{some multiple of } b \in \text{span}(b);$$

STEP 3: $x_2 = \text{mult. of } x_1 + \text{mul. of } \nabla f(x_1) + \text{mult. of } x_0 + \dots \in \text{span}(Ab_1, b)$;

STEP 4: $x_3 = \text{mult.of } x_2 + \dots \nabla f(x_2) + \text{previous iterates} \dots$

$$Ax_2 - b = A \cdot (sAb + tb) - b = sA^2b + tAb - b \in \text{span}(b, Ab, A^2b), \text{ where } s \text{ and } t \text{ are scalars};$$

STEP 5: $x_4 \in \text{span}(b, Ab, A^2b, A^3b)$.

Notice that we can make some linear combinations of the vectors we have available and $A(A(sb + tAb) + uAb + vb) + e(sb + tAb) + fAb + gb \in \text{span}(b, Ab, A^2b, A^3b)$.

Idea: first compute the basis of $\text{span}(b, Ab, A^2b, A^3b)$, then look for the best solution inside this subspace.

The following family of algorithms “uses” the matrix A only by computing matrix-vector products.

Observation 12.3.1. The cost of multiplying a sparse matrix A with a vector z is $O(\text{nnz}(A))$, where $\text{nnz}(A)$ is the number of non zero entries of A .

Proof. Let us assume the matrix A is stored as a vector, whose entries are (i, j, A_{ij}) .

The matrix-vector product would then be

1. `x=zeros`
2. `for (i, j, Aij) such that Aij ≠ 0`
3. $x_i = x_i + A_{ij} * z_j$
4. `end`

□

From now on we will consider to have a function `compute_product_with_A` that we use to compute matrix-vector products with A . In particular, $x = \text{compute_product_with_A}(z)$ computes $x = Az$, given z . This function will be the only way in which the matrix A appears in our algorithm. If A is sparse, hence, these algorithms become particularly fast. Moreover, if we somehow have matrices that are not really sparse, but for which there exists a clever implementation of the matrix-vector product, this class of algorithms will give good results.

Definition 12.3.1 (Krylov subspace). *Let $A \in M(m, \mathbb{R})$ and let $b \in \mathbb{R}^m$. The Krylov subspace of index n is $\mathbf{K}_n(\mathbf{A}, \mathbf{b}) = \text{span}(b, Ab, A^2b, \dots, A^{n-1}b)$.*

Equivalently, $k \in K_n(A, b) \iff \exists \alpha_1, \dots, \alpha_{n-1} \in \mathbb{R}^m$ s.t. $v = \alpha_0 b + \alpha_1 Ab + \alpha_2 A^2b + \dots + \alpha_{n-1} A^{n-1}b$.

Equivalently, $(\alpha_0 + \alpha_1 A + \alpha_2 A^2 + \dots + \alpha_{n-1} A^{n-1})b = p(A)b$ for a polynomial p of degree such that $\deg(p) \leq n - 1$.

Observation 12.3.2 (Properties).

1. $v, w \in K_n(A, b) \Rightarrow \alpha + \beta W \in K_n(A, b);$
2. $v \in K_n(A, b) \Rightarrow Av \in K_n(A, b)$. *Proof.* Let us take $v = \alpha_0 b + \dots + \alpha_{n-1} b$, then $Av = A(\alpha_0 b + \dots + \alpha_{n-1} b) = \alpha_0 Ab + \dots + \alpha_{n-1} A^n b;$
3. $\dim(K_n(A, b)) \leq n$. It is exactly n if $b, Ab, A^2b, \dots, A^{n-1}b$ are linearly independent.
4. Let us assume $\dim(K_n(A, b)) \leq n$. In the second point, if A^{n-1} was really necessary $\alpha_{n-1} \neq 0$ or $v \in K_n(A, b)$, $v \notin K_{n-1}(A, b)$, equivalently then $A^n b$ is really necessary to write Av , i.e. $Av \in K_{n+1}(A, b)$ but $Av \notin K_n(A, b)$.
5. We may observe that $\dim(K_1(A, b)) < \dim(K_2(A, b)) < \dots < \dim(K_{n_{\max}}(A, b)) = \dim(K_{n_{\max}+1}(A, b)) = \dots$

Lecture 13

29th of November 2018 — F. Poloni

In this lecture we address the problem of designing methods that use Krylov spaces to solve linear systems.

We need alternative methods to Gaussian elimination because matrices are too large.

13.0.1 Naive idea

Find first a “good” search subspace, then look for best approximation of the solution of $Ax = b$ in that space.

Let us assume that our space is the image of a matrix V $Im(V)$. $\forall x \in Im(V)$ such x may be written as $x = V^1y_1 + V^2y_2 + \dots + V^ny_n$, where V^i are the columns of the matrix V .

The idea is to find a vector y that satisfies $\min_{y \in \mathbb{R}} \|AVy - b\|$, which is a least squares problem.

13.0.2 Improvement

A good search space is $Im(V) = K_n(A, b) = \text{span}(b, Ab, A^2b, \dots, A^{n-1}b)$, where $K_n(A, b)$ is the Krylov space.

The issue here is that $V = (b \quad Ab \quad \dots \quad A^{n-1}b)$ is a bad basis, because it is ill conditioned.

Working with that V is problematic: its columns “tend” to be aligned with the dominant eigenvector and this means that V is close to a rank 1 matrix.

We need a better basis for $K_n(A, b)$, in particular an orthogonal basis.

💡 Do you recall?

An orthogonal basis is a basis in which each couple of vectors are orthogonal

13.1 Arnoldi algorithm

The idea behind this algorithm is to build an orthogonal basis of $K_n(A, b)$ incrementally.

The algorithm at a generic step, takes an orthogonal basis for $K_n(A, b)$ and adds a vector to produce one of $K_{n+1}(A, b)$.

Let us assume that we start with $\{q_1, q_2, \dots, q_n\}$, orthogonal basis of $K_n(A, b)$.

We also assume that $q_n = \alpha_0 b + \alpha_1 A b + \dots + \alpha_{n-1} A^{n-1} b = p(A)b$, where we impose $\alpha_{n-1} \neq 0$.

$p(A) = \alpha_0 I + \alpha_1 A + \dots + \alpha_{n-1} A^{n-1}$ has degree exactly 1.

STEP 1: $K_1(A, b) = \text{span}(b)$, $q_1 = \frac{q_1}{\|q_1\|}$

GENERIC STEP:

1. produce a vector $K_{n+1}(A, b) - K_n(A, b) : w = Aq_n$
2. $w \in K_{n+1}(A, b) = 0$ $w = q_1\beta_1 + q_2\beta_2 + \dots + q_n\beta_n + q_{n+1}\beta_{n+1}$, where $q_1, q_2, \dots, q_n, q_{n+1}$ is an orthogonal basis of $K_{n+1}(A, b)$. In this context q_1, q_2, \dots, q_n are known, while q_{n+1} is still to be determined.

Notice that $\forall i = 1, \dots, n$ holds the following $q_i^T w = q_i^T q_1\beta_1 + \dots + q_i^T q_n\beta_n + q_i^T q_{n+1}\beta_{n+1} = q_i^T q_i\beta_i = \beta_i$, because of the orthonormality of the basis. Now we can compute $q_{n+1}\beta_{n+1} = w - q_1\beta_1 - q_2\beta_2 - \dots - q_n\beta_n = z$. If we choose $\beta_{n+1} = \|z\|$ we have that $q_{n+1} = \frac{z}{\|z\|}$ and this produces a valid choice of q_{n+1} . We still need to prove that it has norm 1 and it's orthogonal to all the other vectors in the basis.

An implementation of this algorithm is shown in Algorithm 13.1.1.

ALGORITHM 13.1.1 Arnoldi algorithm Matlab implementation.

```

1  function Q = arnoldi(A, b, n)
2  Q = zeros(length(b), n); %will be filled in
3  H = zeros(n+1, m);
4  Q(:, 1) = b / norm(b);
5  for j = 1 : n
6    w = A * Q(:, j);
7    for i = 1:j
8      % not what we showed earlier here, but stabler
9      betai = Q(:, i)' * w;
10     w = w - betai * Q(:, i);
11     H(i, j) = betai;
12   end
13   nrm = norm(w);
14   H(j+1, j) = nrm;
15   Q(:, j+1) = w / nrm;
16 end

```

Notice that we presented an algorithm where $\beta_i = q_i^T w$ for $i = 1, \dots, n$, then $w \leftarrow w - \beta_1 q_1 - \beta_2 q_2 - \dots - \beta_n q_n$.

In the implementation we compute $\beta_i = q_1^T w$, $w \leftarrow w - \beta_1 q_1$, $\beta_2 = q_2^T w$, $w \leftarrow w - \beta_2 q_2$. Why? It is more stable.

At step j

$$Aq_j = \beta_{1,j}q_1 + \beta_{2,j}q_2 + \dots + \beta_{j,j}q_j + \beta_{j+1,j}q_{j+1} = Q \begin{bmatrix} \beta_{1,j} \\ \beta_{2,j} \\ \vdots \\ \beta_{j+1,j} \\ 0 \\ 0 \\ \vdots \\ 0 \end{bmatrix}.$$

$$Q = \begin{bmatrix} q_1 & q_2 & \cdots & q_n & q_{n+1} \end{bmatrix}$$

$$AQ = A \begin{bmatrix} q_1 & q_2 & \cdots & q_n & q_n \end{bmatrix} = Q \begin{bmatrix} \beta_{1,1} & \beta_{1,2} & \beta_{1,3} & \cdots & \beta_{1,n} \\ \beta_{2,1} & \beta_{2,2} & \beta_{2,3} & \cdots & \beta_{2,n} \\ 0 & \beta_{3,2} & \beta_{3,3} & \cdots & \beta_{3,n} \\ 0 & 0 & \beta_{4,3} & \cdots & \beta_{4,n} \\ \vdots & \ddots & \ddots & \ddots & \vdots \\ 0 & 0 & \cdots & 0 & \beta_{n+1,n} \end{bmatrix}$$

Hence the values q_i , $\beta_{i,j}$ computed by Arnoldi satisfy $AQ_n = Q_{n+1}H$, where $H \in M(n+1, n, \mathbb{R})$ and looks like a triangular matrix plus a diagonal below,

namely $H = \begin{bmatrix} \beta_{1,1} & \beta_{1,2} & \beta_{1,3} & \cdots & \beta_{1,n} \\ \beta_{2,1} & \beta_{2,2} & \beta_{2,3} & \cdots & \beta_{2,n} \\ 0 & \beta_{3,2} & \beta_{3,3} & \cdots & \beta_{3,n} \\ 0 & 0 & \beta_{4,3} & \cdots & \beta_{4,n} \\ \vdots & & \ddots & \ddots & \vdots \\ 0 & 0 & \cdots & 0 & \beta_{n+1,n} \end{bmatrix}$

Notation: $Q_{n+1} = \begin{bmatrix} q_1 & q_2 & \cdots & q_n & q_{n+1} \end{bmatrix}$, $Q_n = \begin{bmatrix} q_1 & q_2 & \cdots & q_n \end{bmatrix}$
such that $Q_n \in M(m, n, \mathbb{R})$, while $Q_{n+1} \in M(m, n+1, \mathbb{R})$, $A \in M(m, \mathbb{R})$ and $b \in M(m, n)$.

FIGURE 13.1: We started from a matirx A which has many entries and it gets factorized by the product of two smaller matrices

For every matrix A there exist and Arnoldi factorization.

We don't like that we multiply A by two different matrices on the left and on the right, but we may improve the algorithm by using the same matrix.

$AQ_n = (Q_n|q_{n+1}) \cdot \text{matriToDraw} = Q_nH + q_{n+1} \cdot (0, \dots, 0, *) = Q_nH + q_{n+1} \cdot \text{tre}_{n+1}\beta_{j+1,j}$, where $e_i = 0 \dots 010 \dots 0$.

Last remark: $AQ_n = Q_{n+1}H$ doesn't allow a factorization of the matrix A , because Q_n isn't invertible, because Q_n is tall, thin and it doesn't have an inverse.

Once we see how Arnoldi works, we would like to see how to use it.

At some point this assumption must become false. For instance, assume we arrive at step $m = n$: q_1, q_2, \dots, q_m are a basis of \mathbb{R}^m , so

$$Aq_m = \beta_1 q_1 + \cdots + \beta_m q_m + 0$$

(without an additional term $\beta_{m+1}q_{m+1}$.)

Arnoldi factorization for $m = n$ (assuming nothing broke down before):

$AQ_m = Q_mH$ is a factorization into square matrices, formally $Q_m, H, A \in M(m, \mathbb{R})$, so we can write something that we couldn't write before:

$A = A_m H Q_m^T$ we recall that H is upper triangular plus another diagonal before.

Definition 13.1.1. Let $H \in M(m, \mathbb{R})$ such that $H_{ij} = 0 \forall i > j + 1$. H is called **Hessemberg matrix**.

Fact 13.1.1. *The QR factorization an Hessemberg matrix $H \in M(m, \mathbb{R})$ can be computed in $O(m^2)$ operations.*

This approach may be used, but in practice, since A is large and sparse we do not want to go until the end.

Let us analyze what happens if at some point $K_{n+1}(A, b) = K_n(A, b)$?

$$Aq_n = \beta_1 q_1 + \cdots + \beta_n q_n + 0$$

for instance, if b is an eigenvector of A , it happens already at $n = 1$.

In the implementation Algorithm 13.1.1 if we have a breakdown at step $n + 1$ this means that $w = Aq_n$ has already enough q_i s, so $q_{n+1}\beta_{n+1} = 0$.

Problem: $q_{n+1} = \frac{z}{\|z\|}$ division by 0. We need to change the definition $\beta_{n+1} = \|z\| = 0$. At this point we don't get a basis of the Krylov space, but we can still go on as "nothing happened", as long as these vectors are orthonormal. We go on until the end we get

$$AQ_m = H_m Q_m,$$

$$H_m = \left[\begin{array}{cccc|cccccc} * & \cdots & * & * & * & \cdots & \cdots & * \\ * & \cdots & * & * & * & \cdots & \cdots & * \\ 0 & \ddots & * & * & \vdots & \vdots & \vdots & \vdots \\ 0 & 0 & * & * & * & \cdots & \cdots & * \\ \hline & & & 0 & * & \cdots & * & * \\ & & & & * & \cdots & * & * \\ & & & & 0 & \ddots & * & * \\ & & & & 0 & 0 & * & * \end{array} \right].$$

The blocks are square.

This is good news, because it allows us to make a lot of manipulations.

$$A = Q_m H_m Q_m^T = \begin{bmatrix} Q_n & \hat{Q} \end{bmatrix} \begin{bmatrix} H_n & L \\ 0 & M \end{bmatrix} \begin{bmatrix} Q_n & \hat{Q} \end{bmatrix}^T$$

Do you recall?

We said two things about block triangular matrices:

1. The eigenvalues of the matrix are the eigenvalues of the diagonal blocks, more formally, $eigh(A) = eigh(H) = eigh(H_n) \cup eigh(M)$
2. We can solve linear systems on block matrices more easily. In this case,

$$\begin{aligned}
x = A^{-1}b &= Q_m H^{-1} Q_m^T b = Q_m H^{-1} \begin{bmatrix} q_1^T \\ q_2^T \\ \vdots \\ q_m^T \end{bmatrix} b \stackrel{(1)}{=} Q_m H^{-1} \begin{bmatrix} \|b\| \\ 0 \\ \vdots \\ 0 \end{bmatrix} \\
&= \begin{bmatrix} Q_n & \hat{Q} \end{bmatrix} \begin{bmatrix} H_n^{-1} & -H_n^{-1}LM^{-1} \\ 0 & M^{-1} \end{bmatrix} \begin{bmatrix} \|b\| \mathbf{e}_1 \\ 0 \end{bmatrix} \\
&= \begin{bmatrix} Q_n & \hat{Q} \end{bmatrix} \begin{bmatrix} \|b\| H_n^{-1} \mathbf{e}_1 \\ 0 \end{bmatrix} = Q_n \|b\| H_n^{-1} \mathbf{e}_1.
\end{aligned} \tag{13.1.1}$$

where $\stackrel{(1)}{=}$ follows from the fact that $b = \|b\| q_1$ and q_1 is orthogonal to all the other q_i .

An attentive reader may notice that at step j , when encountering $B_{j+1,j} = 0$, we know the following:

- some eigenvalues of A (those of H_n)
- Given an eigencouple v, λ such that $H_n v = \lambda v$ then $Q_n v$ is an eigenvector of A with eigenvalue λ
- the solution $x = Q_n H_n^{-1} \|b\| e_1$ of $Ax = b$ and this is called lucky breakdown.

The point is that we have what we need to compute the solution at last step before the breakdown.

Theorem 13.1.2. ‘Lucky breakdown’: if it happens at an early step, we can solve linear systems (or compute some eigenvalues) cheaply: costs n matrix-vector products + $O(mn^2)$.

What happens when there is no breakdown? After n steps of Arnoldi,

1. if $H_n v = \lambda v$ is an eigenpair of H_n . $Q_n v, \lambda$ is an approximation of an eigenpair of A .
2. $\tilde{x} = Q_n H_n^{-1} \|b\| e_1$ is an approximantion of the solution x of $Ax = b$

Lecture 14

5th of December 2018 — F.
Poloni

💡 Do you recall?

Arnoldi factorization: we wanted to build an orthonormal base $\{q_1, q_2, \dots, q_n\}$ of the Krylov space $K_n(A, b) = \{b, Ab, A^2b, \dots, A^{n-1}b\}$. We showed that the matrix A could be more or less factorized as $AQ_n = Q_{n+1}H_n = Q_nH_n + q_{n+1}h_{n+1,n}e_{n+1}^T$.

Moreover, we concluded that we could approximate the eigenvalues of the matrix A through the eigenvalues of matrix H_n , while the eigenvectors are obtained as Q_nv , where v is an eigenvector of the matrix H_n .

In the first part of the lecture we run some experiments on MatLab and we observed that the eigenvalues are approximated better and better starting from the boundaries. Notice that the approximation is not always good, but it is good enough if we take into account the complexity of the algorithm.

We are interested in explaining the convergence of Arnoldi method.

14.1 Convergence of Arnoldi

The eigenvalues of H_n are eigenvalues of a “nearby matrix” obtained by taking \tilde{H}_m (result of the full process) and replacing $\tilde{H}_{n+1,n}$ with zero.

$$\tilde{H}_m = \left[\begin{array}{cccc|cccc} * & \cdots & * & * & * & \cdots & \cdots & * \\ * & \cdots & * & * & * & \cdots & \cdots & * \\ 0 & \ddots & * & * & \vdots & \vdots & \vdots & \vdots \\ 0 & 0 & * & * & * & \cdots & \cdots & * \\ \hline & & \textcircled{0} & & * & \cdots & * & * \\ & & & & * & \cdots & * & * \\ & & & & 0 & \ddots & * & * \\ & & & & 0 & 0 & * & * \end{array} \right] \rightarrow H_m = \left[\begin{array}{cccc|cccc} * & \cdots & * & * & * & \cdots & \cdots & * \\ * & \cdots & * & * & * & \cdots & \cdots & * \\ 0 & \ddots & * & * & \vdots & \vdots & \vdots & \vdots \\ 0 & 0 & * & * & * & \cdots & \cdots & * \\ \hline & & \textcircled{*} & & * & \cdots & * & * \\ & & & & * & \cdots & * & * \\ & & & & 0 & \ddots & * & * \\ & & & & 0 & 0 & * & * \end{array} \right]$$

We expect that this change does not lead to a significative change in the eigenvalues, in other words, the eigenvalues of \tilde{H}_m differ from the eigenvalues of H_m by $|h_{n+1,n}|$. Formally, $\|\tilde{H}_m - H_m\| = h_{n+1,n}$.

14.1.1 Better explanation

The space $K_n(A, b) = \text{span}(b, Ab, \dots, A^{n-1}b)$ contains the right “features” to represent the eigenvectors of A with largest eigenvalues: if $A = V\Lambda V^{-1}$ is diagonalizable, then

$$\begin{aligned} A^k b &= (V\Lambda V^{-1}) \cdots (V\Lambda V^{-1}) b \\ &= V\Lambda^k V^{-1} b \\ &= \begin{pmatrix} V^1 & V^2 & \cdots & V^m \end{pmatrix} \cdot \begin{pmatrix} \lambda_1^n & & & \\ & \lambda_2^n & & \\ & & \ddots & \\ & & & \lambda_m^n \end{pmatrix} \begin{pmatrix} c_1 \\ c_2 \\ \vdots \\ c_m \end{pmatrix} \quad (14.1.1) \\ &= V^1 \lambda_1^k c_1 + V^2 \lambda_2^k c_2 + \cdots + V^m \lambda_m^k c_m \end{aligned}$$

where $c = V^{-1}b$.

$A^k b$ is a linear combination of the eigenvectors V^i in which those with largest $|\lambda_i|$ are “more prominent”, in other words as n increases the components involving the largest $|\lambda_i|$ s grow faster.

This also tells us that $\text{span}(V^1, V^2, \dots, V^m)$ (that are the eigenvectors associated to largest eigenvalues in modulus) “represent well” $K_m(A, b) = \text{span}(b, Ab, \dots, A^{m-1}b)$.

We cannot provide a formal proof of the convergence, because there are some counter examples.

Example 14.1.1. Let $A \in M(m, \mathbb{R})$, such that $A = \begin{pmatrix} 0 & & & 1 \\ 1 & 0 & & \\ & 1 & \ddots & \\ & & \ddots & \\ & & & 1 & 0 \end{pmatrix}$

In this case the eigenvalues are 0 until the last iteration. In the last step the eigenvalues become correct.

Notice that in this case the absolute values of the eigenvalues are the same, hence we are not able to do the trick explained above.

Observation 14.1.1 (Matlab syntax). The command $[V, D] = \text{eigs}(A)$ does not work on sparse matrices A . In this case we may run Arnoldi method and use the best values obtained by Arnoldi: $[V, D] = \text{eigs}(A, n)$, which computes approximations of the top- n (largest in modulus) eigenvalues.

Notice that the Matlab “implementation” (the quotes are because the Arnoldi method de facto is not implemented in Matlab) of the Arnoldi method uses some tricks to converge fast.

It is possible to use the command eigs and pass to it a λ -function which computes the matrix-vector product and this is useful when the matrix and the vector have a particular shape.

We would like to understand how to compute some eigenvalues that are not the biggest.

Lemma 14.1.1. Let $A \in M(m, \mathbb{R})$ and let $(\lambda_1, \mathbf{v}_1), \dots, (\lambda_k, \mathbf{v}_k)$ be the eigenvalues/vectors of A . The following holds:

1. $(\lambda_i + \alpha, \mathbf{v}_i)$ are eigenvalues/vectors of $A + \alpha I$;
2. $(\frac{1}{\lambda_i}, \mathbf{v}_i)$ are eigenvalues/vectors of A^{-1} ;
3. $(\lambda_i^k, \mathbf{v}_i)$ are eigenvalues/vectors of A^k .

Proof. Let us omit the subscript i to ease notation:

1. $(A + \alpha I)\mathbf{v} = A\mathbf{v} + \alpha\mathbf{v} = \lambda\mathbf{v} + \alpha\mathbf{v} = (\lambda + \alpha)\mathbf{v}$;
2. $(\lambda^{-1}\mathbf{v})$ is an eigenpair of A^{-1} . We need to check that $A^{-1}\mathbf{v} = \lambda^{-1}\mathbf{v}$. If we multiply by λA both sides: $\lambda A A^{-1}\mathbf{v} = \lambda A \lambda^{-1}\mathbf{v} \Leftrightarrow \lambda\mathbf{v} = \lambda\lambda^{-1}A\mathbf{v}$, which is true by definition of eigenvalue/vector of A ;
3. (by induction) $A^2\mathbf{v} = A(A\mathbf{v}) = A \cdot \lambda\mathbf{v} = \lambda A\mathbf{v} = \lambda\lambda\mathbf{v} = \lambda^2\mathbf{v}$.

□

This lemma gives us the chance to use Arnoldi algorithm to compute the eigenvalues of a slightly modified matrix $B = (A - \mu I)^{-1}$. If (λ, v) is an eigenpair of A , then $(\lambda - \mu^{-1}, v)$ is an eigenpair of B .

When is that $(\lambda - \mu^{-1})$ is large? When λ and μ are close.

If A has eigenvalues $\lambda_1, \lambda_2, \dots, \lambda_m$, then the eigenvalues of B are $\nu_1 = \frac{1}{\lambda_1 - \mu}, \nu_2 = \frac{1}{\lambda_2 - \mu}, \dots, \nu_m = \frac{1}{\lambda_m - \mu}$.

The problem is that such matrix B is not sparse when A is sparse.

We perform a trick to overcome this issue: we pass the function `eig` a λ -function that computes the product $Bz = A - \mu I^{-1}z$ without computing B explicitly. The idea is to use factorizations: $A - \mu I = LU$, then $Bz = U^{-1}(L^{-1}z)$, that we can compute by back-substitution.

```
% computes 5 eigenvalues closest to mu=2
» f1 = eigs(A, 5, mu);
» f1 = eigs(A, 5, 'SM'); %smallest magnitude
» f1 = eigs(A, 5, 'LM'); %largest magnitude
```

Notice that `eigs(A, 6, 1)` needs to factorize $A - 1 \cdot I = LU$, which might induce a lot of fill-in.

As final observation, the equivalents to Matlab's `eigs` function are `scipy.linalg.eigs` for Python and `arpack` for C/C++ and Fortran.

Lecture 15

7th of December 2018 — F. Poloni

In this lecture we are interested in using Arnoldi method to solve linear systems.

We can use the *sparse eigenvalues function* that we saw in last lecture and the **Generalized Minimum RESidual** (GMRES).

Our task is to approximate the solution of a large-scale linear system of the form $Ax = b$ and our approach is to look for “the closest thing to solution” inside $K_n(A, b)$.

Through Arnoldi of A , b and n , we obtained $[Q, H]$ and we can approximate the solution x as $Q_1y_1 + Q_2y_2 + \dots + Q_ny_n = Qy$, which is a good approximation of the solution inside $K_n(A, b)$, formally

$$\min_{x \in K_n(A, b)} \|Ax - b\|, \quad x = Q_n y.$$

which is equivalent to $\min_{y \in \mathbb{R}^n} \|AQ_n y - b\|$.

We can perform some more reductions and:

$$\begin{aligned} \|AQ_n y - b\| &\stackrel{(1)}{=} \|Q_{n+1}\underline{H}_n y - b\| \\ &\stackrel{(2)}{=} \|Q_{n+1}\underline{H}_n y - Q_{n+1}\|b\|e_1\| \\ &= \|Q_{n+1} \cdot (\underline{H}_n y - \|b\|e_1)\| \\ &\stackrel{(3)}{=} \|\underline{H}_n y - \|b\|e_1\|. \end{aligned} \tag{15.0.1}$$

where $\stackrel{(1)}{=}$ is due to the equivalence $AQ_n = Q_{n+1}H_n$, with $H_n \in M(n+1, n)$, $\stackrel{(2)}{=}$ follows from the fact that $q_1 = \frac{b}{\|b\|}$ and $\stackrel{(3)}{=}$ is explained recalling that Q_{n+1} is an orthogonal rectangular matrix in $M(mn+1)$ and $\|z\| = \|Q_{n+1}z\|$, since $z^T z = z^T Q_{n+1}^T Q_{n+1} z$.

We got a LS problem of size $(n+1) \times n$ (small), where $y \in \mathbb{R}^n$ and $e_1 \in \mathbb{R}^{n+1}$; moreover \underline{H} has the following shape

$$H = \left[\begin{array}{cc|cc} * & \cdots & * & * & * & \cdots & \cdots & * \\ * & \cdots & * & * & * & \cdots & \cdots & * \\ 0 & \ddots & * & * & \vdots & \vdots & \vdots & \vdots \\ 0 & 0 & * & * & * & \cdots & \cdots & * \\ \hline & & & \textcircled{*} & * & \cdots & * & * \\ 0 & & & & * & \cdots & * & * \\ & & & & 0 & \ddots & * & * \\ & & & & 0 & 0 & * & * \end{array} \right]$$

hence it is quite sparse.

$gr(H)$ can be computed in $O(n^2)$ using the fact that H is ‘almost triangular’ (Hessenberg matrix), although it is not a big optimization, since n Arnoldi steps need to be computed first.

Notice that instead of doing a QR at the end, we can compute QRs of H_1, H_2, \dots and update them at each step. This allows us to compute at each step a residual $\|Ax_n - b\|$ that we can use as stopping criterion.



Something on Matlab ...

Matlab has `gmres(A, b)` (and Python has `scipy.sparse.linalg.gmres`).

To estimate the convergence of GMRES we can see x as a polynomial ($x = p(A)b$, such that $p(t) = \alpha_0 + \alpha_1 t + \cdots + \alpha_{n-1} t^{n-1}$ is a polynomial of degree $n-1$).

As far as the residual is concerned $Ax - b = A \cdot p(A) \cdot b - b = A \cdot (\alpha_0 I + \alpha_1 A + \cdots + \alpha_{n-1} A^{n-1}) \cdot b - b = q(A) \cdot b$, where $q(t) = t \cdot p(t) - 1$. If $A = V \Lambda V^{-1}$ diagonalizable, then $A^k = V \cdot \begin{bmatrix} \lambda_1^k & & \\ & \ddots & \\ & & \lambda_m^k \end{bmatrix} V^{-1}$ and

$$q(A) = V \begin{bmatrix} q(\lambda_1) & & \\ & \ddots & \\ & & q(\lambda_m) \end{bmatrix} V^{-1}.$$

All this computation was needed to write the residual GMRES in a clearer form:

$$\begin{aligned}
\min_{x \in K_n(A, b)} \|Ax - b\| &= \min_{\substack{q(x) = xp(x)-1 \\ \text{of degree } \leq n}} \|Ap(A)b - b\| \\
&= \min_{\substack{q(x) = xp(x)-1 \\ \text{of degree } \leq n}} \|q(A)b\| \\
&\leq (\min_{\dots} \|q(A)\|) \cdot \|b\| \\
&= \min_{\dots} \left\| V \begin{bmatrix} q(\lambda_1) & & \\ & \ddots & \\ & & q(\lambda_m) \end{bmatrix} V^{-1} \right\| \quad (15.0.2) \\
&\leq \min_{\dots} \|V\| \cdot \left\| \begin{bmatrix} q(\lambda_1) & & \\ & \ddots & \\ & & q(\lambda_m) \end{bmatrix} \right\| \cdot \|V^{-1}\| \\
&\leq K(V) \cdot \left\| \min_{\dots} \begin{bmatrix} q(\lambda_1) & & \\ & \ddots & \\ & & q(\lambda_m) \end{bmatrix} \right\|
\end{aligned}$$

If A has very few distinct eigenvalues ($k \leq n$ of them), then we can find q such that $q(\lambda_i) = 0$ for all i and $q(0) = -1$, hence n steps of GMRES give us the exact solution.

If A has eigenvalues clustered in n points in the plane, we can find a polynomial q such that $q(\lambda_i)$ is small for all i .

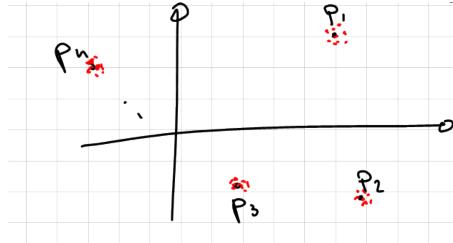


FIGURE 15.1: In this picture the eigenvalues are clustered around P_1, P_2, P_3 and P_4 . We can find a polynomial q such that $q(\text{red points}) \approx 0$.

Notice that Gauss operations on the rows of any matrix A (e.g. swapping rows or scalar multiplication of a row) change its eigenvalues, without changing the solution.

More generally, given $P \in M(n, \mathbb{R})$ we can change the problem $Ax = b$ to $PAx = Pb$. If P is invertible, the two systems have the same solution. However, the spectrum of PA may be much better (in the above sense) than the spectrum of A , leading to a faster solution with GMRES.

In particular, this happens if we manage to find $P \approx A^{-1}$. The perfect choice would be $P = A^{-1}$, but, of course, if we knew A^{-1} we would already have a way to solve linear systems: just compute the matrix multiplication $A^{-1}b$.

There are various techniques (often problem-dependent) to build effective *preconditioners* P . One comes from approximate LU factorizations of A (in a suitable sense); they are known as *incomplete LU* preconditioners.

Lecture 16

13th of December 2018 —
F. Poloni

16.1 Lanczos algorithm

💡 Do you recall?

In last lecture we saw how to factorize a matrix $A \in M(m, \mathbb{R})$ with Arnoldi (i.e. $AQ_n = Q_{n+1}\underline{H}_n = Q_n\underline{H}_n + h_{n+1,n}q_{n+1}e_1^T$).

If A is **symmetric**, something special happens: $\underline{H}_n = Q_m^T A Q_m$ is symmetric as well, so it is a **tridiagonal** matrix. This improves the complexity of the Arnoldi process, because many iterations of the for loop ($j = 1 : n$) are not needed anymore, we need only two iterations.

This symmetric variant of Arnoldi is called *Lanczos algorithm*, and such algorithm reduces the cost to n matrix products + $O(mn)$.

Suppose $A = A^T$ is positive definite. Then, we can find the solution to $Ax = b$ by minimizing the (strictly convex) function $f(x) = \frac{1}{2}x^T Ax - b^T x$.

Surprisingly, conjugate gradient on this problem can be interpreted as a Krylov subspace method.

The pseudocode of such algorithm can be found in Algorithm 16.1.1, where x_k is the current iterate, $r_k = b - Ax_k = -\nabla f(x_k)$ is the residual and d_k is the search direction.

ALGORITHM 16.1.1 Pseudocode for the conjugate gradient method.

```

1: procedure CG_ITERATION
2:    $x_0 \leftarrow 0;$ 
3:    $r_0 \leftarrow b;$ 
4:    $d_0 \leftarrow b;$ 
5:   for  $k = 1:n$  do
6:      $\alpha_k \leftarrow (r_{k-1}^T r_{k-1}) / (d_{k-1}^T A d_{k-1});$ 
7:      $x_k \leftarrow x_{k-1} + \alpha_k d_{k-1};$ 
8:      $r_k \leftarrow r_{k-1} - \alpha_k A d_{k-1};$ 
9:      $\beta_k \leftarrow (r_k^T r_k) / (r_{k-1}^T r_{k-1});$ 
10:     $d_k \leftarrow r_k + \beta_k d_{k-1};$ 
11:   end for
12: end procedure

```

Notice that the search direction (line 10) is modified adding a multiple of the previous search direction to the residual and β_k is chosen such that d_k and d_{k-1} are A -orthogonal (formally, $d_k^T A d_{k-1} = 0$).

Conversely, the next point is chosen in order to minimize the objective function $f(x_{k-1} \alpha_k d_{k-1})$.

As far as the complexity is concerned, the space complexity is constant (three vectors) and the time complexity is $O(mn)$.

Theorem 16.1.1. $K_k(A, b) = \text{span}(x_1, x_2, \dots, x_k) = \text{span}(d_0, d_1, \dots, d_{k-1}) = \text{span}(r_0, r_1, \dots, r_{k-1})$.

Theorem 16.1.2. *The residuals are orthogonal and the search directions are A -orthogonal. Formally, $r_j^T r_k = d_i^T A d_k = 0, \forall i < k$.*

Proof. By induction: Let us assume we proved the thesis $r_j^T r_k = 0$ for $k-1, k-2, \dots, 0$. Since $x_k = x_{k-1} \alpha_k d_{k-1}$, the residual $r_k = b - Ax_k = b - A(x_{k-1} + \alpha_k d_{k-1}) = b - Ax_{k-1} - \alpha_k A d_{k-1} = r_{k-1} - \alpha_k A d_{k-1}$.

Let us compute $r_j^T r_k = r_j^T \cdot (r_{k-1} - \alpha_k A d_{k-1})$.

- If $i < k-1$ then $r_j^T r_{k-1} - r_j^T \alpha_k A d_{k-1} = 0$, because the first term is 0 from induction hypothesis and $r_j^T \alpha_k A d_{k-1} = 0$, because $r_j \in K_{k-1}(A, b) = \text{span}(d_0, d_1, \dots, d_{k-2})$.
- If $i = k-1$, $0 = r_{k-1}^T \cdot (r_{k-1} - \alpha_k A d_{k-1}) = r_{k-1}^T r_{k-1} - \alpha_k r_{k-1}^T A d_{k-1}$ holds if $\alpha_k = \frac{r_{k-1}^T r_{k-1}}{r_{k-1}^T A d_{k-1}}$. We are left with proving that $\alpha_k = \frac{r_{k-1}^T r_{k-1}}{r_{k-1}^T A d_{k-1}} = \alpha_k = \frac{r_{k-1}^T r_{k-1}}{d_{k-1}^T A d_{k-1}}$.

This is true, since $d_{k-1} = r_{k-1} + \beta_{k-1} d_{k-2}$, so $d_{k-1}^T A d_{k-1} = (r_{k-1} + \beta_{k-1} d_{k-2})^T A d_{k-1} = r_{k-1}^T A d_{k-1} + \beta_{k-1} d_{k-2}^T A d_{k-1}$ and the second part is equal to 0 by induction.

□

Notice that this base is not orthonormal, we need to rescale it to obtain an orthonormal one, moreover, $\frac{1}{\|r_i\|} r_i$ coincides (up to a sign) with the q_i obtained with Arnoldi.

We are left with writing the equation we need to solve at each iteration, namely we need to ensure that $r_k = b - Ax_k$ is orthogonal to all vectors of $K_k(A, b)$ which is equivalent to requiring $Q_k^T \cdot (b - Ax_k) = 0$ or, equivalently, $\|b\| \cdot e_1 = H_k y_k$.

In figure Figure 16.1 we can see a comparison between Arnoldi algorithm and the conjugate gradient.

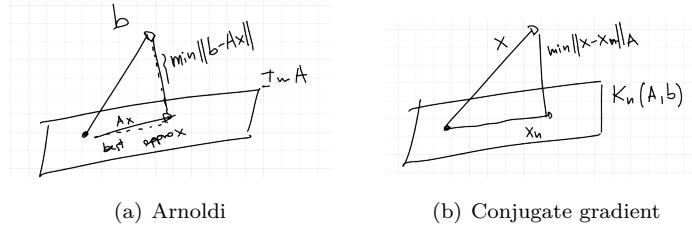


FIGURE 16.1: Traditional orthogonality (Arnoldi) leads to the minimization of the 2-norm, while in the conjugate gradient we impose A -orthogonality and we get a good approximation in several norms.

As far as convergence speed is concerned,

Theorem 16.1.3. x_k is the best approximation of the exact (and unknown) solution x to $Ax = b$ in $K_k(A, b)$ in the A -norm, i.e. $x_k = \arg \min_{z \in K_k(A, b)} (x - z)^T A (x - z)$

Theorem 16.1.4. Let λ_{\max} , λ_{\min} be the maximum/minimum eigenvalue of A ; then, CG converges with rate

$$\|x - x_k\| \leq \left(\frac{\sqrt{\lambda_{\max}} - \sqrt{\lambda_{\min}}}{\sqrt{\lambda_{\max}} + \sqrt{\lambda_{\min}}} \right)^k \|x - x_0\|.$$

We can rewrite it in terms of a more familiar quantity: for a positive definite matrix, eigenvalues and singular values coincide, hence

$$\frac{\sqrt{\lambda_{\max}} - \sqrt{\lambda_{\min}}}{\sqrt{\lambda_{\max}} + \sqrt{\lambda_{\min}}} = \frac{\sqrt{\sigma_1} - \sqrt{\sigma_m}}{\sqrt{\sigma_1} + \sqrt{\sigma_m}} = \frac{\sqrt{\frac{\sigma_1}{\sigma_m}} - 1}{\sqrt{\frac{\sigma_1}{\sigma_m}} + 1} = \frac{\sqrt{\kappa(A)} - 1}{\sqrt{\kappa(A)} + 1}.$$

For large values of $\kappa(A)$, this is approximately $1 - \frac{2}{\sqrt{\kappa(A)}}$, while if $\kappa(A) \approx 1$ the convergence speed is very high.

As for GMRES, if A has only n different eigenvalues, then this minimum reaches 0 after n steps. If the eigenvalues of A are ‘clustered’, one can construct polynomials such that $q(\lambda)$ is small for each λ then fast convergence is implied.

Lecture 17

19th of September 2018

A. Frangioni

This course will deal with the optimization and numerical analysis of machine learning problems. We are not going to discuss difficult problems (e.g. NP-hard problems), besides we will present an efficient solution for simple ones (often **convex** ones), since we are dealing with huge amount of data.

Let us start with a warm up on machine learning problems.

17.1 Introduction to machine learning problems

Machine learning techniques are not as “young” as it might seem, the intuition has been there for ages, but we did not have enough calculus power. Machine learning algorithms have started working well recently, thanks to the many improvements in computer performances; for this reason, it is becoming a more and more popular subject to study.

The main idea behind machine learning is to take a huge amount of data (e.g. frames of a video for object-recognition) and squeeze them, in order to process them. This intuitive concept is translated into mathematical terms as “building a **model**” that fits our data. As in practical engineering problems, people want to construct a model (a small sized representation of the large thing we want to produce in the end) and try to understand its behaviour, before actually build such an object. Take as an example the problem of designing a jet. It is not clever to start building the plane before designing a cheap prototype to better study its behaviour in the atmosphere.

The kind of models we want to build are cheap to construct and as close as possible to the real problem we are studying. In physics, people try to find the best mathematical model to describe a real world phenomenon. The main issue is computation, since the more accurate the model, the more costly the prediction phase. Hence, a good model is a trade-off between accuracy and simplicity, namely it provides good prediction without incurring in slow computations.

The model, though, has to be parametric: we do not have only one model, we have a “shape” of a model, which is fit to our problem through the tuning of some parameters.

Example 17.1.1. As an example, we are given three couples: $f(x_1) = y_1$, $f(x_2) = y_2$, $f(x_3) = y_3$, as shown in Figure 17.1.

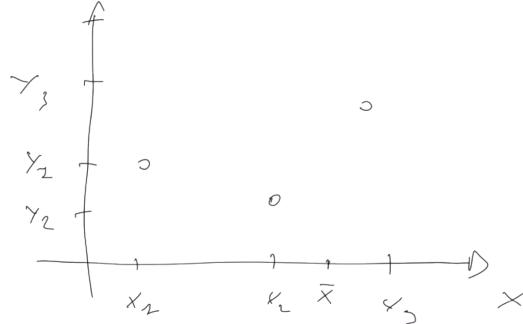


FIGURE 17.1: Geometric representation of the input. We are interested in finding a model that fits the input data and allows to predict \bar{y} out of \bar{x} .

We need to make some choices: first, we need to decide the kind of model we believe is a good approximation of the objective function, say a linear model $f(x) = ax + b$. After doing that, we are left with choosing its parameters (in order to pick a line among the whole family of linear functions), namely a and b .

The aim is to build a model that fits the data we are given and then tune the parameters in order to achieve a good accuracy for a given application (recall that the model is parametric, hence the right values for the parameters have to be learned).

Another important characteristic of a good model is that it should not take too long to be built.

In this course we do not concentrate on the problem of finding the model that best fits our data, but we are already given a problem and a model and we only study its behaviour through its parameters. In other words, within the family of models with a given outline, we want to find the one that better represents the phenomena observed. This is called **fitting** and it is clearly some sort of optimization problem, where the fitting task is typically the computational bottleneck.

However, machine learning is more than simply fitting: fitting minimizes training error (or empirical risk), but ML aims at minimizing test error (i.e. generalization error).

A machine learning solution first builds a model that fits the observed data and then performs an hyper-parameter tuning in order to achieve a good “predicting power” on unseen inputs. In machine learning, a model that fits the

data while achieving good performances on new examples is said to be “non **overfitting**”.

For machine learning purposes, a mathematical model should be:

- *accurate* (describing well the process under consideration)
- *computationally inexpensive* (providing answers rapidly)
- *general* (it can be applied to many different processes)

It goes without saying that achieving all these goals is practically impossible.

17.2 Optimization

In the rest of this lecture we are going to better understand what an optimization problem is, through some intuitive real world examples.

◆ Terminology

In this course we will refer to vectors using the bold notation $\mathbf{v} \in \mathbb{R}^n$. The i -th entry of vector \mathbf{v} is denoted as $v_i \in \mathbb{R}$, while the j -th pattern in the training set is a couple of vectors $(\mathbf{x}^j, \mathbf{y}^j)$.

For more details on vectors and vector spaces see Section 18.4.

17.2.1 Linear estimation

Let us consider a phenomenon measured by one number $y \in \mathbb{R}$ that is believed to depend on a vector $\mathbf{x} = [x_1, \dots, x_n]$. We are provided a set of observations: $\{(y^1, \mathbf{x}^1), \dots, (y^m, \mathbf{x}^m)\}$.

Definition 17.2.1 (Linear model). *Let $f : \mathbb{R}^n \rightarrow \mathbb{R}^n$ be the objective function. We call $\tilde{f}(\mathbf{x}) = \sum_{i=1}^n w_i x_i + w_0 = \mathbf{w}_+^T \mathbf{x} + w_0$ the **linear model** of f for a given set of parameters, which is a vector $\mathbf{w} = (w_0, \mathbf{w}_+) = (w_0, w_1, \dots, w_n) \in \mathbb{R}^{n+1}$.*

How can we evaluate the “similarity” between our model and the objective function? Through computing the “error” or difference between the objective function value and the model prediction on each input. Under this assumption, the error function may be used to find the best parameters for our model, through solving a minimum problem:

Definition 17.2.2 (Least squares problem). *Let $f : \mathbb{R}^n \rightarrow \mathbb{R}^n$ be the objective function, such that $f(\mathbf{x}) = \mathbf{y}$ and let $X\mathbf{w}$ be our linear model. Then we can find the best values for vector $\mathbf{w} \in \mathbb{R}^{n+1}$ by solving*

$$\min_{\mathbf{w}} \|\mathbf{y} - X\mathbf{w}\|$$

$$\text{where } Y = \begin{pmatrix} \mathbf{y}^1 \\ \vdots \\ \mathbf{y}^m \end{pmatrix} \text{ and } X = \begin{pmatrix} 1 & \mathbf{x}^1 \\ \vdots & \vdots \\ 1 & \mathbf{x}^m \end{pmatrix}.$$

If the matrix X is invertible then the simple solution is $\mathbf{w} = X^{-1}\mathbf{y}$. The point is that this operation is very costly when dealing with a huge number of entries (in the next paragraph we will see a way to manage it).

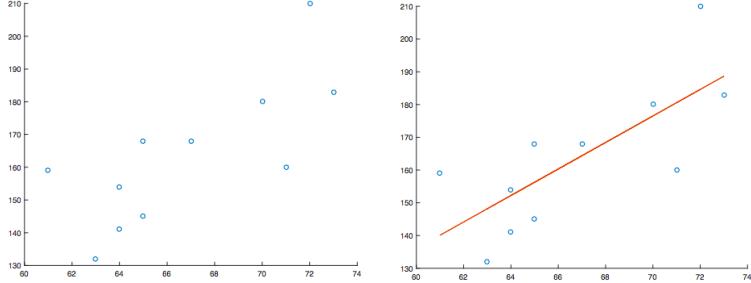


FIGURE 17.2: A linear estimation fitting example

17.2.2 Low-rank approximation

A (large, sparse) matrix $M \in M(n, m, \mathbb{R})$ describes a phenomenon depending on pairs (e.g., objects bought by customers) and we may want to approximate that matrix as the product between two smaller matrices (find a few features that describe most of users' choices): a “tall thin” matrix $A \in M(n, k, \mathbb{R})$ and a “fat large” $B \in M(k, m, \mathbb{R})$ (where $k \ll n, m$).

$$\boxed{M} \approx \boxed{A} \cdot \boxed{B}$$

This problem can be translated into a numerical analysis problem of the following shape

$$\min_{A, B} \|M - AB\|$$

The matrices A and B can be obtained from eigenvectors of $M^T M$ and MM^T , but that's a huge, possibly dense matrix. So a more efficient way should be used, which also avoids the explicit formation of $M^T M$ and MM^T because that would need a lot of memory.

Efficiently solving this problem requires:

- low-complexity computation
- avoiding the explicit forming of $M^T M$ and MM^T
- exploiting the structure of M (sparsity, similar columns, ...)
- ensuring that the solution is *numerically stable*

17.2.3 Support vector machines

Let us consider the so-called “decision problem”: given a set of values of many parameters (variables) “label” a person i as ill or healthy, $y^i \in \{1, -1\}$.

The geometric intuition in two dimensions is given by Figure 17.3. We would like to find the line that better splits the plane into two regions, because this could help to diagnose the next patient. The rationale here is to maximize the space between the line and the nearest points (called **margin**), in order to have a better accuracy.

The distance between the two hyperplanes (w_0, \mathbf{w}_+) and (w'_0, \mathbf{w}_+) in Figure 17.3 is computed as $\frac{2}{\|\mathbf{w}_+\|}$, so in order to maximize the distance between the planes we aim at minimizing $\|\mathbf{w}_+\|$. For each hyperplane that lies between (w_0, \mathbf{w}_+) and (w'_0, \mathbf{w}_+) the following holds

$$\begin{cases} \mathbf{w}_+ \mathbf{x}^i + w_0 \geq 1 & \text{if } y^i = 1 \\ \mathbf{w}_+ \mathbf{x}^i + w_0 \leq -1 & \text{if } y^i = -1 \end{cases}$$

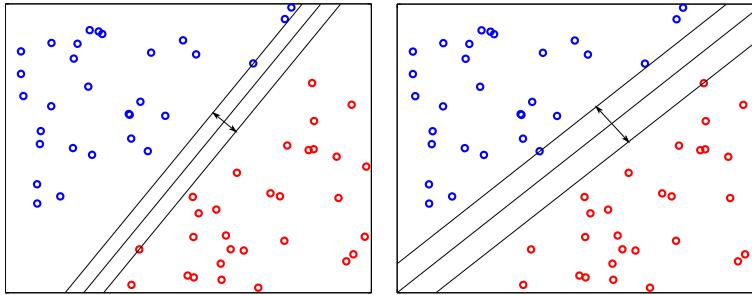


FIGURE 17.3: There are many possible boundaries that can be chosen as a model using many angular coefficients. Our best guess is the one that maximizes the distance between the line and the nearest points.

Under the hypotheses of this optimization problem, the maximum-margin separating hyperplane (assuming that any exists) is the solution of

$$\min_{\mathbf{w}} \{\|\mathbf{w}_+\|^2 : y^i(\mathbf{w}_+ \mathbf{x}^i + w_0) \geq 1, i = 1, \dots, m\}$$

In practice, most of the times there is no such line. To overcome this issue we introduce the concept of “penalty” that accounts for the number of points that are misclassified, through the following

Definition 17.2.3 (SVM Primal problem). *We term Support Vector Machine primal problem (SVM-P) a convex constrained problem with complex constraints that is formalized as*

$$\begin{cases} \min_{\mathbf{w}, \xi} \|\mathbf{w}_+\|^2 + C \sum_{i=1}^m \xi_i \\ y^i(\mathbf{w}_+ \mathbf{x}^i + w_0) \geq 1 - \xi_i, \xi_i \geq 0, \forall i = 1, \dots, m \end{cases} \quad (\text{SVM-P})$$

Where C is an **hyper-parameter** that weights the violation of the separating margin.

This definition formalizes the intuition that the approximated function may have a greater norm and lead to a very small misclassification error, or it could be the other way round. Both these solutions are acceptable and their performances depend only on the problem.

Whenever we are able to solve a multi-objective optimization problem we are also able to solve what is called the **dual problem**, which in our case has the following shape.

Definition 17.2.4 (SVM Dual problem). *We call Support Vector Machine dual problem (SVM-D) a convex constrained quadratic problem defined as*

$$\begin{cases} \max_{\alpha} \sum_{i=1}^m \alpha_i - \frac{1}{2} \sum_{i=1}^m \sum_{j=1}^m \alpha_i < \mathbf{x}^i, \mathbf{x}^j > \alpha_j \\ \sum_{i=1}^m y^i \alpha_i = 0 \\ 0 \leq \alpha_i \leq C, \forall i = 1, \dots, m \end{cases} \quad (\text{SVM-D})$$

The idea behind the theory of duality is to obtain the solution of a problem by solving an (apparently) different one.

Lemma 17.2.1. *Given an optimal solution α^* for the dual problem (SVM-D), then $\mathbf{w}^* = \sum_{i=1}^m \alpha_i^* y^i \mathbf{x}^i$ is optimal for the primal problem (SVM-P).*

A reader who has a deeper background in the field of machine learning knows that the scalar product in the dual form allows the usage of the so-called “kernel trick”. This mathematical transformation allows the mapping of the input space into a larger feature space where the points are more likely to be linearly separable.

Theoretically, the feature space can be infinite-dimensional, provided that the scalar product can be (efficiently) computed.

This whole course has the aim of presenting some techniques for solving efficiently **convex quadratic problems**, as the ones presented above.

Lecture 18

21st of September 2018

A. Frangioni

18.1 Mathematical background for optimization problems

Definition 18.1.1 (Minimum problem). *Let S be a set, called **feasible region** and let $f : S \rightarrow \mathbb{R}$ be any function, called **objective function**. We term **minimum problem** the problem of finding the minimum value of such f . Formally,*

$$f_* = \min_x \{f(x) : x \in S\} \quad (\text{P})$$

Definition 18.1.2 (Feasible solution). *Let $x \in F$ be a solution of the minimum problem in which the domain is a superset of $S \subset F$. We say that x is a **feasible solution** if $x \in S$. Conversely, x is **unfeasible** if $x \in F \setminus S$.*

Definition 18.1.3 (Optimal solution). *Under the same hypothesis of the above definition, we call x_* that realizes $f(x_*) = f_*$ an **optimal solution**, where $f_* \leq f(x) \forall x \in S, \forall v > f_* \exists x \in S$ s.t. $f(x) < v$.*

It is possible to find problems where there is no optimal solution at all.

Observation 18.1.1. *There are two cases in which it is not possible to find an optimal solution:*

1. *The domain is empty, which may be not trivial to prove, since it is an NP-hard problem sometimes;*
2. *We want to find the minimum of the objective function but the function is unbounded below ($\forall M \exists x_M \in S$ s.t. $f(x_M) \leq M$). Conversely, we aim to maximize a function, which is unbounded above.*

Example 18.1.1 (Bad optimization problems). *The following practical cases are examples of problems that do not admit an optimal solution:*

- *empty case ($S = \emptyset$):* $\min\{f(x) = x : x \in \mathbb{R} \wedge x \leq -1 \wedge x \geq 1\}$
- *unbounded (below):* $\min\{f(x) = x : x \in \mathbb{R} \wedge x \leq 0\}$
- *bad f and S:* $\min\{f(x) = x : x \in \mathbb{R} \wedge x > 0\}$
- *bad f:* let us consider an iterative algorithm that moves towards the optimum. It may happen that the function decreases and increases along a certain direction, such function is not continuous so it is impossible to reach the optimum:

$$\min \left\{ f(x) = \begin{cases} x & \text{if } x > 0 \\ 1 & \text{if } x = 0 \end{cases} \quad x \in [0, 1] \right\}$$

Solving an optimization problem can be one of the following procedures:

1. Finding x_* and proving it is optimal
2. Proving $S = \emptyset$
3. Constructively proving that the function is unbounded ($\forall M \exists x_M \in S$ s.t. $f(x_M) \leq M$).

Typically “ $x \in \mathbb{R}$ ” actually means “ $x \in \mathbb{Q}$ ” with up to k digits precision and most of the times we consider optimal a solution which is close to the true optimal value, modulo some error (we call \bar{x} , the approximated optimal).

Definition 18.1.4 (Absolute error). We call **absolute error** the gap between the real value and the one we obtained. Formally,

$$f(\bar{x}) - f_* \leq \varepsilon$$

Definition 18.1.5 (Relative error). We term **relative error** the absolute error, normalized by the true value of the function

$$(f(\bar{x}) - f_*) / |f_*| \leq \varepsilon$$

18.1.1 Multi-objective Optimization

It may happen that there are more than one function that need to be minimized (maximized) and they could be contrasting or have incomparable units (apples vs oranges).

In multi-objective optimization, typically there is not a feasible solution that minimizes all objective functions simultaneously. It is in this context that *Pareto optimal solutions* are introduced: such solutions cannot be improved in any of the objectives without degrading at least one of the other objectives.

Definition 18.1.6 (Pareto frontier). Let (P) be the minimum multi-objective optimization problem formalized as

$$\min_x \{[f_1(x), \dots, f_k(x)] : x \in S\}$$

A feasible solution $x^1 \in S$ is said to (**Pareto**) dominate another solution $x^2 \in S$, if

$$\begin{cases} f_i(x^1) \leq f_i(x^2) & \forall i \in \{1, 2, \dots, k\} \\ f_j(x^1) \leq f_j(x^2) & \forall j \in \{1, 2, \dots, k\} \end{cases}$$

A solution $x^* \in S$ (and the corresponding outcome $f(x^*)$) is called **Pareto optimal**, if there does not exist another solution that dominates it. The set of Pareto optimal outcomes is often called the **Pareto frontier**.

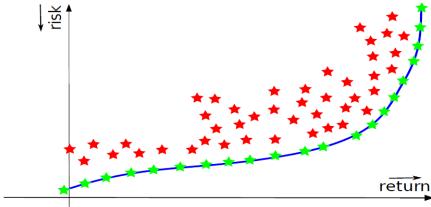


FIGURE 18.1: An example of Pareto frontier

Example 18.1.2. Let us take two functions f_1 and f_2 . We want to solve the minimization problem

$$\min_x \{[f_1(x), f_2(x)] : x \in S\} \quad (P)$$

and to do so we present two different approaches:

SCALARIZATION. Using a linear combination of the two functions, formally, $f(x) = \alpha f_1(x) + \beta f_2(x)$. An example, where $\alpha = 1$ is shown in Figure 18.2.

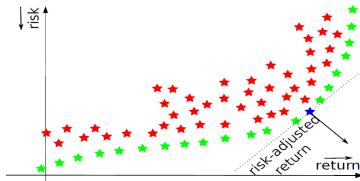


FIGURE 18.2: Maximize risk-adjusted return

BUDGETING. Intuitively corresponds to taking into account only one objective function and add the others as constraints, provided that the values of the other functions are not too high. In formal terms, $f(x) = f_1(x)$, where $S := S \cup \{x \in S : f_2(x) \leq b\}$. As an example see Figure 18.3.

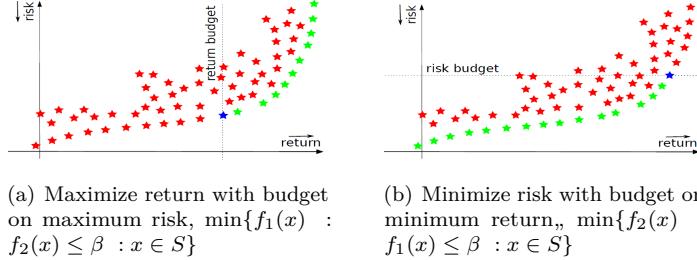


FIGURE 18.3: Budgeting

18.2 Infima, suprema and extended reals

Let us introduce some mathematical background on minimization (maximization).

Definition 18.2.1 (Totally ordered set). *We say that a set S is **totally ordered** if $\forall x, y \in S$, either $f(x) \leq f(y)$ or $f(y) \leq f(x)$.*

Definition 18.2.2 (Infima and suprema). *Let R be a totally ordered set and let s be one of its subsets ($S \subseteq R$):*

\underline{s} is the **infimum** of S ($\underline{s} = \inf S$) iff $\underline{s} \leq s \quad \forall s \in S \quad \wedge \quad \forall t > \underline{s} \exists s \in S$ s.t. $s \leq t$

\bar{s} is the **supremum** of S ($\bar{s} = \sup S$) iff $\bar{s} \geq s \quad \forall s \in S \quad \wedge \quad \forall t < \bar{s} \exists s \in S$ s.t. $s \geq t$

An attentive reader may notice that it is possible that infima (suprema) do not exist in \mathbb{R} .

Definition 18.2.3 (Extended real). *In the case of unbounded functions the value of infima or suprema are ∞ , and we call **extended reals** $\overline{\mathbb{R}} = -\infty \cup \mathbb{R} \cup +\infty$.*

Property 18.2.1. *The following holds for the extended reals:*

- For all $S \subseteq \mathbb{R}$, $\sup/\inf S \in \overline{\mathbb{R}}$
- $\inf S = -\infty$ indicates that there is no (finite) infimum
- $\inf \emptyset = \infty$, $\sup \emptyset = -\infty$

18.3 (Monotone) Sequences in \mathbb{R} and optimization

We are interested in studying sequences, because iterative methods start from a certain point and move towards the optimal.

❖ Terminology

We denote a sequence of iterates as $\{x_i\} \subset S$ and we denote the plugging a function f into such a sequence as $v_i = f(x_i)$.

Definition 18.3.1 (Limit). *Given a sequence $\{x_i\}$ the **limit** for $i \rightarrow \infty$ is defined as*

$$\lim_{i \rightarrow \infty} v_i = v \iff \forall \varepsilon > 0 \exists h \text{ s.t. } |v_i - v| \leq \varepsilon \forall i \geq h$$

It may happen that a sequence has or does not have a limit. For example $\{\frac{1}{n}\}$ has limit 0 for $n \rightarrow +\infty$, while $\{(-1)^n\}$ does not have any.

Fact 18.3.1. *Let us be given a monotone sequence, then the sequence **does** have a limit.*

Notice that a sequence either is monotone or it can be “split” into two monotone sequences.

Example 18.3.1. *Let us consider the sequence $\{(-1)^n\}$. It does not converge to any value, but it can be split into $\{(-1)^{2n}\}$ and $\{(-1)^{2n+1}\}$ and these two subsequences are both monotone.*

Forcing monotonicity on sequences in \mathbb{R} can be performed splitting the sequence into a non-decreasing sequence and a non-increasing sequence:

$$\underline{v}_i^* = \inf\{v_h : h \leq i\} \text{ and } \bar{v}_i^* = \sup\{v_h : h \leq i\}$$

This way we get $\underline{v}_1 \leq \underline{v}_2 \leq \dots$ and $\bar{v}_1 \geq \bar{v}_2 \geq \dots$ and they have a limit: $\liminf_{i \rightarrow \infty} \underline{v}_i \lim_{i \rightarrow \infty} \underline{v}_i^* = \underline{v}_\infty^* = f_*$ and $\limsup_{i \rightarrow \infty} \bar{v}_i \lim_{i \rightarrow \infty} \bar{v}_i^* = \bar{v}_\infty^* = f_*$

Fact 18.3.2. *The following holds:*

- $\bar{v}_i \geq \underline{v}_i \Rightarrow \limsup_{i \rightarrow \infty} v_i \geq \liminf_{i \rightarrow \infty} v_i$
- $\lim_{i \rightarrow \infty} v_i = v \iff \limsup_{i \rightarrow \infty} v_i = v = \liminf_{i \rightarrow \infty} v_i$

18.4 Vector spaces and topology

For the problems we are going to study, real numbers are used to describe each feature of a dataset and this has the mathematical equivalent in the term “vector”.

Definition 18.4.1 (Euclidean vector space). *We call **Euclidean space***

$$\mathbb{R}^n := \left\{ \begin{pmatrix} x_1 \\ x_2 \\ \vdots \\ x_n \end{pmatrix} : x_i \in \mathbb{R}, i = 1, \dots, n \right\}$$

Equivalently, we can characterize the Euclidean space as Cartesian product of \mathbb{R} n times: $\mathbb{R}^n = \mathbb{R} \times \mathbb{R} \times \dots \mathbb{R}$. Every vector space and in particular Euclidean spaces are closed under sum and scalar multiplication.

The main operations on elements of the Euclidean space (vectors) are:

$$\text{SUM: } \mathbf{x} + \mathbf{y} := \begin{pmatrix} x_1 + y_1 \\ x_2 + y_2 \\ \vdots \\ x_n + y_n \end{pmatrix}$$

$$\text{SCALAR MULTIPLICATION: } \alpha \mathbf{x} = \begin{pmatrix} \alpha x_1 \\ \alpha x_2 \\ \vdots \\ \alpha x_n \end{pmatrix}$$

In linear algebra, $\mathbf{x} \in \mathbb{R}^n$ is a column vector $\in \mathbb{R}^n$, whenever a row vector is needed for dimensionality issues it is denoted as x^T .

Definition 18.4.2 (Basis). *Any vector $\mathbf{x} \in \mathbb{R}^n$ can be obtained from a linear combination of a set of vectors that is called “basis”. In Euclidean spaces there exists a special basis that is called “canonical”, where each vector \mathbf{u}^i has a 1 in position i and a 0 elsewhere.*

Definition 18.4.3 (Finite vector space). *Let $V(K, n, m)$ be a vector space. It is said to be **finite** if its basis have a finite cardinality.*

At this point, it is crucial to observe that not all vector spaces are finite nor they are a totally ordered set.

In this course we will deal with the concept of limit over a vector space very often and for this purpose we are going to introduce a topology on vector spaces (i.e., norm, scalar product, distance).

Definition 18.4.4 (Euclidean norm). Let $\mathbf{x} \in \mathbb{R}^n$, we define **euclidean norm** of \mathbf{x} :

$$\|\mathbf{x}\|_2 = \sqrt{\sum_{i=1}^n x_i^2} = \sqrt{\langle \mathbf{x}, \mathbf{x} \rangle}$$

Fact 18.4.1. Any norm on a vector space has the following properties:

1. $\|\mathbf{x}\| \geq 0$ and $\forall \mathbf{x} \in \mathbb{R}^n$, $\|\mathbf{x}\| = 0 \iff \mathbf{x} = \mathbf{0}$;
2. $\|\alpha \mathbf{x}\| = |\alpha| \|\mathbf{x}\|$, $\forall \mathbf{x} \in \mathbb{R}^n$, $\alpha \in \mathbb{R}$;
3. $\|\mathbf{x} + \mathbf{y}\| \leq \|\mathbf{x}\| + \|\mathbf{y}\|$, $\forall \mathbf{x}, \mathbf{y} \in \mathbb{R}^n$ (triangle inequality).

Fact 18.4.2 (Cauchy-Schwartz inequality). Let $\mathbf{x}, \mathbf{y} \in \mathbb{R}^n$. The following holds:

$$\langle \mathbf{x}, \mathbf{y} \rangle^2 \leq \langle \mathbf{x}, \mathbf{x} \rangle \langle \mathbf{y}, \mathbf{y} \rangle \equiv |\langle \mathbf{x}, \mathbf{y} \rangle| \leq \|\mathbf{x}\| \|\mathbf{y}\|, \quad \forall \mathbf{x}, \mathbf{y} \in \mathbb{R}^n$$

Fact 18.4.3 (Parallelogram Law).

$$2\|\mathbf{x}\|^2 + 2\|\mathbf{y}\|^2 = \|\mathbf{x} + \mathbf{y}\|^2 + \|\mathbf{x} - \mathbf{y}\|^2$$

Fact 18.4.4.

$$\|\mathbf{x} + \mathbf{y}\|^2 = \|\mathbf{x}\|^2 + \|\mathbf{y}\|^2 + 2\langle \mathbf{x}, \mathbf{y} \rangle$$

The 2-norm is a special case of the more general p -norm.

Definition 18.4.5 (p -norm). Let $p \geq 1$ be a real number, the p -norm of a vector $\mathbf{x} = (x_1, \dots, x_n)$ is defined as follow:

$$\|\mathbf{x}\|_p := \left(\sum_{i=1}^n |x_i|^p \right)^{1/p}$$

We require $p \geq 1$ for the general definition of the p -norm because the triangle inequality fails to hold if $p < 1$. The p -norm is convex for $p \geq 1$ and non convex for $p < 1$.

The most important p -norms are the following:

- $\|\mathbf{x}\|_1 := \sum_{i=1}^n |x_i|$
- $\|\mathbf{x}\|_\infty := \max\{|x_i| : i = 1, \dots, n\}$
- $\|\mathbf{x}\|_i := |\{i : |x_i| > 0\}|$

Fact 18.4.5. For any given finite-dimensional vector space $V(K, n)$ (e.g. \mathbb{R}^n), all norms on V are equivalent (therefore, convergence in one norm implies convergence in any other norm). Formally, $\forall \|\cdot\|_A, \|\cdot\|_B$:

$$\exists 0 < \alpha < \beta \text{ s.t. } \alpha \|x\|_A \leq \|x\|_B \leq \beta \|x\|_A \quad \forall \mathbf{x} \in V$$

This rule may not apply in infinite-dimensional vector spaces such as function spaces. For such cases we define the

Fact 18.4.6 (Holder's inequality). *Let $V(K, n)$ be a (possibly infinite) vector space, for any two norms $\|\cdot\|_A, \|\cdot\|_B$ the **Holder's inequality** holds*

$$\langle x, y \rangle^2 \leq \|x\|_A \|y\|_B \quad 1/A + 1/B = 1$$

Definition 18.4.6 (Ball). *Let $\bar{\mathbf{x}} \in \mathbb{R}^n$. We term **ball** centered in $\bar{\mathbf{x}}$ and having ε as radius as the set of points that are close enough to \mathbf{x} : $B(\bar{\mathbf{x}}, \varepsilon) = \{\mathbf{x} \in \mathbb{R}^n : \|\mathbf{x} - \bar{\mathbf{x}}\| \leq \varepsilon\}$.*

Let us take a unit ball, if the center of the unit-ball is in the origin $(0, 0)$, then each point on the unit-ball will have the same p -norm (i.e. 1). The unit ball therefore describes all points that have “distance” 1 from the origin, where “distance” is measured by the p -norm. In Figure 18.4 we may observe the different shapes of the same ball varying the value of p in the p -norm.

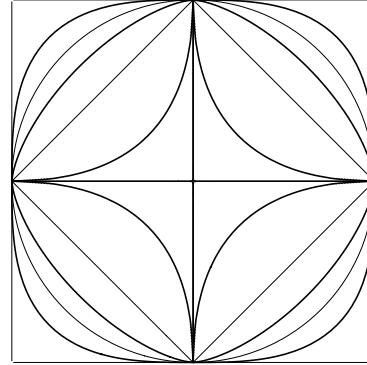


FIGURE 18.4: The shapes of balls centered in the origin of radius 1 varying the value of p -norm.

Definition 18.4.7 (Standard scalar product). *Let $\mathbf{x}, \mathbf{y} \in \mathbb{R}^n$ we define the **standard scalar product** between these two vectors as*

$$\langle \mathbf{x}, \mathbf{y} \rangle := \mathbf{y}^T \mathbf{x} = \sum_{i=1}^n x_i y_i = x_1 y_1 + \cdots + x_n y_n$$

Fact 18.4.7. *A scalar product has the following properties:*

1. $\langle \mathbf{x}, \mathbf{y} \rangle = \langle \mathbf{y}, \mathbf{x} \rangle \quad \forall \mathbf{x}, \mathbf{y} \in \mathbb{R}^n$ (*symmetry*)
2. $\langle \mathbf{x}, \mathbf{x} \rangle \geq 0, \quad \forall \mathbf{x} \in \mathbb{R}^n, \quad \langle \mathbf{x}, \mathbf{x} \rangle = 0 \iff \mathbf{x} = \mathbf{0};$
3. $\langle \alpha \mathbf{x}, \mathbf{y} \rangle = \alpha \langle \mathbf{x}, \mathbf{y} \rangle, \quad \forall \mathbf{x} \in \mathbb{R}^n, \alpha \in \mathbb{R};$

$$4. \langle \mathbf{x} + \mathbf{y}, \mathbf{z} \rangle = \langle \mathbf{x}, \mathbf{z} \rangle + \langle \mathbf{y}, \mathbf{z} \rangle, \forall \mathbf{x}, \mathbf{y}, \mathbf{z} \in \mathbb{R}^n.$$

An important geometric characterization of the scalar product is the one that uses angles:

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

According to this new definition, we can observe that \mathbf{x} and \mathbf{y} are orthogonal ($\mathbf{x} \perp \mathbf{y}$) iff $\langle \mathbf{x}, \mathbf{y} \rangle = 0$ and in all the other cases they somehow point in the same direction.

We could rewrite the scalar product between \mathbf{x} and \mathbf{y} as $\mathbf{y}^T I \mathbf{x}$, where $I \in D(\mathbb{R}, n)$ is the identity matrix. Thanks to this observation, we can state the more general form of the scalar product:

Definition 18.4.8 (Scalar product). *Let $\mathbf{x}, \mathbf{y} \in \mathbb{R}^n$ and let $M \in M(\mathbb{R}, n)$ be a positive definite matrix (for further details see). We call **scalar product** the following*

$$\langle \mathbf{x}, \mathbf{y} \rangle_M := \mathbf{y}^T M \mathbf{x}$$

Aggiungere reference al punto in cui si definisce una matrice definita positiva

Definition 18.4.9 (Euclidean distance). *The **Euclidean distance** between points \mathbf{x} and \mathbf{y} in \mathbb{R}^n is the length of the line segment connecting them. In Cartesian coordinates, if $\mathbf{x} = (x_1, x_2, \dots, x_n)$ and $\mathbf{y} = (y_1, y_2, \dots, y_n)$ are two points in the n -dimensional Euclidean space, then the distance (d) from \mathbf{x} to \mathbf{y} , or from \mathbf{y} to \mathbf{x} is given by*

$$d(\mathbf{x}, \mathbf{y}) := \|\mathbf{x} - \mathbf{y}\| = \sqrt{(x_1 - y_1)^2 + \dots + (x_n - y_n)^2}$$

Fact 18.4.8. *The Euclidean distance has the following properties:*

1. $d(\mathbf{x}, \mathbf{y}) \geq 0 \quad \forall \mathbf{x}, \mathbf{y} \in \mathbb{R}^n, d(\mathbf{x}, \mathbf{y}) = 0 \iff \mathbf{x} = \mathbf{y}$
2. $d(\alpha \mathbf{x}, 0) = |\alpha| d(\mathbf{x}, 0) \quad \forall \mathbf{x} \in \mathbb{R}^n, \alpha \in \mathbb{R}$
3. $d(\mathbf{x}, \mathbf{y}) \leq d(\mathbf{x}, \mathbf{z}) + d(\mathbf{z}, \mathbf{y}) \quad \forall \mathbf{x}, \mathbf{y}, \mathbf{z} \in \mathbb{R}^n$ (triangle inequality)

18.5 Limit of a sequence in \mathbb{R}^n

We have now all the tools to define the notion of limit of a sequence in \mathbb{R}^n .

Definition 18.5.1 (Limit of a sequence in the Euclidean space). *Let $\{\mathbf{x}_i\} \subset \mathbb{R}^n$ be a sequence in \mathbb{R}^n . We say that $\{\mathbf{x}_i\}$ converges to a **limit** \mathbf{x} for $i \rightarrow +\infty$ and denote $\lim_{i \rightarrow \infty} \mathbf{x}_i = \mathbf{x}$ or $\{\mathbf{x}_i\} \rightarrow \mathbf{x}$ if $\forall \varepsilon > 0 \exists h$ s.t. $d(\mathbf{x}_i, \mathbf{x}) \leq \varepsilon \quad \forall i \geq h$ or, equivalently, $\forall \varepsilon > 0 \exists h$ s.t. $\mathbf{x}_i \in \mathcal{B}(\mathbf{x}, \varepsilon) \quad \forall i \geq h$ or, even $\lim_{i \rightarrow \infty} d(\mathbf{x}_i, \mathbf{x}) = 0$.*

In the definition of limit there is the idea that the points of $\{\mathbf{x}_i\}$ eventually all come arbitrarily close to \mathbf{x} . Moreover, since \mathbb{R}^n is not totally ordered, there is no obvious \liminf nor \limsup .

Lecture 19

27th of September 2018

A. Frangioni

Definition 19.0.1 (Minimizing sequence). *Let $\{x_i\} \in S$ be a sequence and let $f : S \rightarrow \mathbb{R}^n$. We say that $\{x_i\}$ is a **minimizing sequence** if the sequence of function values $\{f(x_i)\}$ tends to the greatest lower bound of f ($f_* = \min\{f(x) : x \in S\}$).*

Example 19.0.1. *As an example, consider the following minimum problems and sequences:*

$$\min\{f(x) = x : x \in \mathbb{R} \wedge x > 0\} \text{ and the sequence } \{x_i = 1/i\}$$

$$\min\{1/x : x \in \mathbb{R} \wedge x > 0\} \text{ and the sequence } \{x_i = i\}$$

In both cases, $\{f(x_i)\} \rightarrow 0$, but it is not an optimum.

We want conditions that ensure $\{f(x_i)\} \rightarrow f_* \Rightarrow \{x_i\} \rightarrow x_* \in S$ optimal solution.

Definition 19.0.2 (Interior and border of a set). *Given $S \subseteq \mathbb{R}^n$, we say that $\mathbf{x} \in \text{int}(S)$ (\mathbf{x} is an **interior point**) if it lies inside a ball contained in S . Formally, $\exists r > 0$ s.t. $\mathcal{B}(\mathbf{x}, r) \subseteq S$.*

Moreover, we term **border points** those $\mathbf{x} \in \partial(S)$ such that all the points in the ball centered in x lie for a half inside the set S and for the other half outside. Formally, $\forall r > 0 \exists \mathbf{y}, \mathbf{z} \in \mathcal{B}(\mathbf{x}, r)$, where $\mathbf{y} \in S \wedge \mathbf{z} \notin S$. Intuitively, a point x lies on the border if, .

Notice that a point on the boundary is not necessarily inside the ball, in that case we talk about **open set** (a set which is identical to its interior: $S = \text{int}(S)$).

Definition 19.0.3 (Closure of a set). *Let $S \subseteq \mathbb{R}^n$ we term **closure** of S the set $\text{cl}(S) = \text{int}(S) \cup \partial S$.*

Definition 19.0.4 (Closed set). We say that a set $S \subseteq \mathbb{R}^n$ is **closed** if it coincides with its closure: $S = \text{cl}(S)$.

Equivalently, a set S is termed **closed** if its complementary $(\mathbb{R}^n \setminus S)$ is open.

It is interesting to notice that all the functions that lead to minimizing sequences and do not converge to an optimum are all defined in open sets.

Notice that there are sets that are both open and closed, for example \mathbb{R}^n .

Definition 19.0.5 (Bounded set). Let $S \subseteq \mathbb{R}^n$. We say that S is **bounded** if $\exists r > 0$ such that $S \subseteq B(0, r)$.

Intuitively, a bounded set does not go to ∞ .

Definition 19.0.6 (Compact set). Let $S \subseteq \mathbb{R}^n$. We term S **compact** if it is both closed and bounded.

Definition 19.0.7 (Accumulation point). Let $\{x_i\} \subseteq S$ be a sequence. We say that x is an **accumulation point** if $\exists \{x_{n_i}\}$ subsequence of $\{x_i\}$ converging to x . Formally, $\{x_{n_i}\} \rightarrow x$ or $\liminf_{i \rightarrow \infty} d(x_i, x) = 0$.

Theorem 19.0.1 (Bolzano-Weierstrass). Let $\{x_i\} \subseteq S$ be a bounded, real sequence. Then it has a converging subsequence.

Fact 19.0.2. Let S be a compact set and let $\{x_i\} \subseteq S$ be a minimizing sequence for the objective function f . The limit of the sequence is a feasible solution.

Proof. According to Bolzano-Weierstrass theorem $\{x_i\}$ has an accumulation point $x \in S$. Since it is minimizing $f(x)$ is feasible. \square

Why did we say *feasible* but not *optimal*? If the function is not continuous (cfr. Figure 19.1) it may happen that the sequence is minimizing, but the limit is not the optimum.

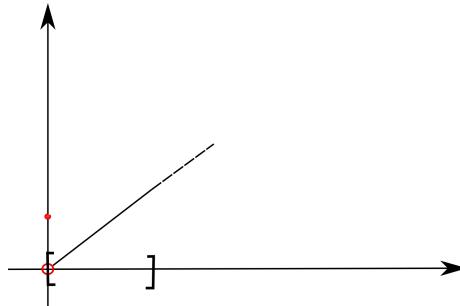


FIGURE 19.1: Case of non-continuity of the objective function in the border point $(0, 0)$.

Definition 19.0.8 (Domain). Let $f : D \rightarrow \mathbb{R}$. We term D **domain** of f and denote $D = \text{dom}(f)$.

In this course we will not take into account the domain of functions, because we can force all functions to be defined in the whole space as follows:

$$f : \mathbb{R}^n \rightarrow \bar{\mathbb{R}}, \text{ where } f(x) = \infty \text{ for } x \notin D.$$

Definition 19.0.9 (Graph and epigraph). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$. We term **graph** $gr(f) = \{(f(\mathbf{x}), \mathbf{x}) : \mathbf{x} \in \text{dom}(f)\}$.

Conversely, we term **epigraph** $\text{epi}(f) = \{(v, \mathbf{x}) : \mathbf{x} \in \text{dom}(f) \wedge v \geq f(\mathbf{x})\}$, see ??.

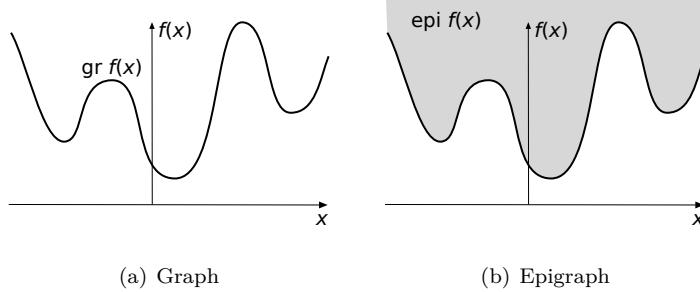


FIGURE 19.2: Graph and epigraph

When dealing with multidimensional inputs spaces it is crucial to have mathematical tools that make possible to somehow “see” the function’s behaviour in a lower-dimensional space.

Definition 19.0.10 (Level and sublevel set). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$. We term **level set** the set of all the inputs that have the same output value. Formally, $L(f, v) = \{\mathbf{x} \in \text{dom}(f) : f(\mathbf{x}) = v\}$.

Conversely, we term **sublevel set** the set of all the inputs which image is smaller than a fixed value v : $S(f, v) = \{\mathbf{x} \in \text{dom}(f) : f(\mathbf{x}) \leq v\}$, see Figure 19.3.

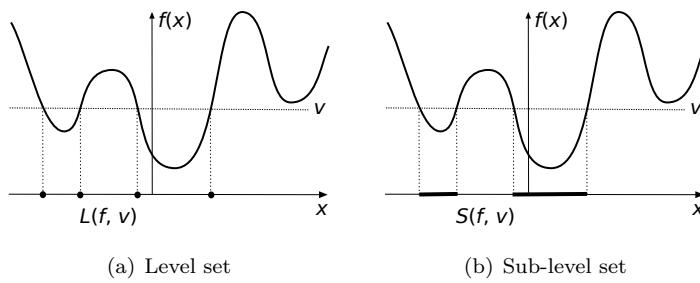


FIGURE 19.3: Level and sub-level sets

19.1 Continuity

Definition 19.1.1 (Continuity). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$. We say that f is **continuous** in \mathbf{x} if $\forall \varepsilon > 0 \exists \delta > 0$ such that $\forall y \in \mathcal{B}(x, \delta) |f(y) - f(x)| < \varepsilon$.

Property 19.1.1. Let $f, g : \mathbb{R}^n \rightarrow \mathbb{R}$. The following holds:

1. $f + g, f \cdot g$ continuous at \mathbf{x}
2. $\max\{f, g\}$ and $\min\{f, g\}$ continuous at \mathbf{x}
3. $f \circ g \equiv f(g(\cdot))$ continuous at \mathbf{x}

Theorem 19.1.2 (Intermediate value). Let $f : \mathbb{R} \rightarrow \mathbb{R}$. f is continuous on $[a, b]$ if $\forall v \in \mathbb{R}$ s.t. $\min\{f(a), f(b)\} \leq v \leq \max\{f(a), f(b)\} \exists c \in [a, b] : f(c) = v$.

Theorem 19.1.3 (Weierstrass extreme value theorem). Let $S \subseteq \mathbb{R}^n$ be a compact set and let f be continuous on S . Then (P) has an optimal solution.

Equivalently, let $S \subseteq \mathbb{R}^n$ compact and let f continuous on S . Then all accumulation points of any minimizing sequence are optima and there is at least one.

Definition 19.1.2 (Lipshitz continuity). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$. We term f **Lipshitz continuous (L.c.)** on $S \subseteq \mathbb{R}^n$ if $\exists L > 0$ such that

$$|f(\mathbf{x}) - f(\mathbf{y})| \leq L \|\mathbf{x} - \mathbf{y}\| \quad \forall \mathbf{x}, \mathbf{y} \in S$$

More generally, f is **globally Lipshitz continuous** if $S = \mathbb{R}^n$ and it is **locally Lipshitz continuous** at \mathbf{x} if $\exists \varepsilon > 0 \exists L > 0 \forall \mathbf{y} \in \mathcal{B}(\mathbf{x}, \varepsilon) |f(\mathbf{x}) - f(\mathbf{y})| \leq L \|\mathbf{x} - \mathbf{y}\|$.

Notice that Lipshitz continuity offers a stronger form of continuity and that the L constant value depends on S . The wider the set the smaller L .

Fact 19.1.4. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$. On a compact set $S \subseteq \mathbb{R}^n$ if f is continuous then it is Lipshitz continuous.

Definition 19.1.3 (Lower(upper) semi-continuity). Let $\{\mathbf{x}_i\} \subseteq \mathbb{R}^n$ be a sequence with accumulation point in \mathbf{x} and let $f : \mathbb{R}^n \rightarrow \mathbb{R}$. f is **lower (upper) semi-continuous (l.(u.)s.c.)** at \mathbf{x} if $f(\mathbf{x}) \leq \liminf_{i \rightarrow \infty} f(\mathbf{x}_i)$ ($f(\mathbf{x}) \geq \limsup_{i \rightarrow \infty} f(\mathbf{x}_i)$).

Equivalently, $\liminf_{\mathbf{y} \rightarrow \mathbf{x}} f(\mathbf{y}) \geq f(\mathbf{x})$, $\limsup_{\mathbf{y} \rightarrow \mathbf{x}} f(\mathbf{y}) \leq f(\mathbf{x})$.

19.2 Derivatives

In this section we address the problem of inferring information on a complicated function around a certain value \bar{x} using very simple functions, that are able to provide reliable information around that \bar{x} . Those functions are called “derivatives”.

Let us start with a brief recap on single-dimension differentiability.

Definition 19.2.1 (Left and right derivative). Let $f : \mathbb{R} \rightarrow \mathbb{R}$. We term **left derivative** $f'_-(x) = \lim_{t \rightarrow 0^-} [f(x+t) - f(x)]/t$.

Conversely, **right derivative** $f'_+(x) = \lim_{t \rightarrow 0^+} [f(x+t) - f(x)]/t$.

Definition 19.2.2 (Differentiable). Let $f : \mathbb{R} \rightarrow \mathbb{R}$. We say that f is **differentiable** at $x \in \text{dom}(f)$ if both left and right derivatives exist and coincide. Formally, $\exists f'_-, f'_+$ and $f'_-(x) = f'_+(x)$.

Fact 19.2.1. Let $f : \mathbb{R} \rightarrow \mathbb{R}$ and differentiable in $x \in \text{dom}(f)$ then f is continuous in x .

Theorem 19.2.2 (Mean value theorem). Let $f : \mathbb{R} \rightarrow \mathbb{R}$ continuous on $[a, b]$ and differentiable on (a, b) then $\exists c \in (a, b)$ s.t. $f'(c) = (f(b) - f(a))/(b - a)$.

Theorem 19.2.3 (Rolle's theorem). Let $f : \mathbb{R} \rightarrow \mathbb{R}$. If $f(a) = f(b)$ then $\exists c \in (a, b)$ s.t. $f'(c) = 0$

Corollary 19.2.4. In the same hypothesis of Rolle's theorem, let a and b consecutive roots of f . Then $f'(a)$ and $f'(b)$ have opposite sign.

19.2.1 Multivariate differentiability

Definition 19.2.3 (Partial derivative). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$. We term **partial derivative** of f w.r.t. x_i at $\mathbf{x} \in \mathbb{R}^n$ as

$$\frac{\partial f}{\partial x_i}(\mathbf{x}) = \lim_{t \rightarrow 0} \frac{f(x_1, \dots, x_{i-1}, x_i + t, x_{i+1}, \dots, x_n) - f(\mathbf{x})}{t}$$

In other words, it is just $f'(x_1, \dots, x_{i-1}, x, x_{i+1}, \dots, x_n)$, considering each component x_j where $j \neq i$ as constant.

Definition 19.2.4 (Gradient). Given $f : \mathbb{R}^n \rightarrow \mathbb{R}$. We term **gradient** of f the vector of all the partial derivatives and we denote $\mathbb{R}^n \ni \nabla f = (\frac{\partial f}{\partial x_1}, \dots, \frac{\partial f}{\partial x_n})^T$.

The gradient is the direction on which the function increases more rapidly, while the opposite of the gradient suggests the direction on which the function decreases most.

Definition 19.2.5 (Directional derivative). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$. We term **directional derivative** of f in point $\mathbf{x} \in \text{dom}(f)$ along direction $\mathbf{d} \in \mathbb{R}^n$

$$\frac{\partial f}{\partial \mathbf{d}}(\mathbf{x}) := \lim_{t \rightarrow 0} \frac{f(\mathbf{x} + t\mathbf{d}) - f(\mathbf{x})}{t}$$

Notice that in a multivariate space it is not possible to assure differentiability checking if the directional derivatives are equal, because there is an infinite number of different directions.

Let us now introduce the notion of multivariate differentiability.

Definition 19.2.6 (Differentiable). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$. We say that f is **differentiable** at $\mathbf{x} \in \mathbb{R}^n$ if \exists a linear function $\phi(h) = \langle c, h \rangle + f(\mathbf{x})$ s.t.

$$\lim_{\|h\| \rightarrow 0} \frac{\|f(\mathbf{x} + h) - \phi(h)\|}{\|h\|} = 0$$

The intuition here is that the linear function should approximate f pretty well around \mathbf{x} .

Fact 19.2.5. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ differentiable at \mathbf{x} . Then f is locally Lipschitz continuous at \mathbf{x} .

Corollary 19.2.6. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ differentiable at \mathbf{x} . Then f is continuous at \mathbf{x} .

Notice that the converse does not hold.

Fact 19.2.7. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$, let $\mathbf{x} \in \mathbb{R}^n$ and let us assume $\exists \delta > 0$ s.t. $\forall i \frac{\partial f}{\partial x_i}(\mathbf{y})$ is continuous $\forall \mathbf{y} \in \mathcal{B}(\mathbf{x}, \delta)$.

Then f differentiable at point \mathbf{x} .

Notice that the converse of Proposition 19.2.7 does not hold either.

We are now ready to introduce a class of functions that allows good results for optimization: \mathcal{C}^1 class.

Definition 19.2.7 (\mathcal{C}^1 class). We call **\mathcal{C}^1 -class** the class of functions with continuous gradient.

Fact 19.2.8. Let $f \in \mathcal{C}^1$, then f is differentiable everywhere and also continuous everywhere.

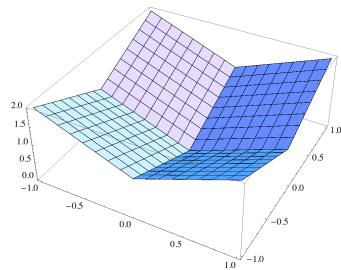
Definition 19.2.8 (First order model). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ and let $\mathbf{x} \in \text{dom}(f)$. We term **first order model** of f at point x

$$L_x(y) = \nabla f(\mathbf{x})(y - \mathbf{x}) + f(\mathbf{x})$$

Fact 19.2.9. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ and let $x \in \text{dom}(f)$ such that f is differentiable at \mathbf{x} . Then $(L_x, f(\mathbf{x})) \perp S(f, f(\mathbf{x})) \perp \nabla f(\mathbf{x})$.

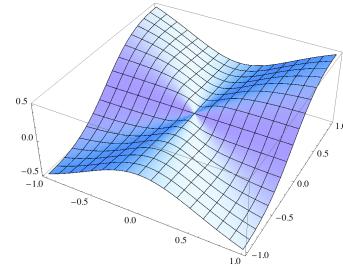
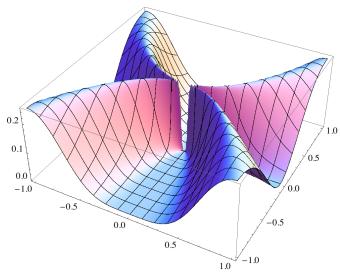
Geometrically speaking, we can observe that a function is smooth whenever its levelsets are smooth.

Example 19.2.1. Let us consider some functions that are not differentiable in the minimum.



The function $f(x_1, x_2) = \| [x_1, x_2] \| = |x_1| + |x_2|$ has some kinks.

The function $f(x_1, x_2) = \frac{x_1^2 x_2}{x_1^2 + x_2^2}$ may be put to 0 in $(0, 0)$ for continuity, but it is still not differentiable in $(0, 0)$ although the function admits directional derivatives in $(0, 0)$.



The function $f(x_1, x_2) = \left(\frac{x_1^2 x_2}{x_1^2 + x_2^2}\right)^2$ is not continuous in 0. There are some directions that lead to the limit 0, while there are some other directions (the parabolas above) that do not lead to value 0.

Lecture 20

3rd of October 2018

A. Frangioni

Example 20.0.1 (On derivatives). Let $f : \mathbb{R}^2 \rightarrow \mathbb{R}$ such that $f(x, y) = x^2 e^y$. Compute the partial derivatives and the directional derivatives in the two directions $\mathbf{d}_1 = (0, 1)^T$ and $\mathbf{d}_2 = (1, 0)^T$.

- $\frac{\partial f}{\partial x} = 2xe^y$
- $\frac{\partial f}{\partial y} = x^2 e^y$
- $\frac{\partial f}{\partial \mathbf{d}_1} = \lim_{t \rightarrow 0} \frac{f(x+t \cdot 0, y+t \cdot 1)}{t} = \lim_{t \rightarrow 0} \frac{f(x, y+t)}{t}$. An attentive reader may notice that the directional derivative in direction \mathbf{d}_1 is equivalent to the derivative on the second component.
- $\frac{\partial f}{\partial \mathbf{d}_2} = \lim_{t \rightarrow 0} \frac{f(x+t \cdot 1, y+t \cdot 0)}{t} = \lim_{t \rightarrow 0} \frac{f(x+t, y)}{t}$. Conversely, with respect to what stated before, the directional derivative of f along the direction \mathbf{d}_2 is equivalent to the partial derivative w.r.t the first component.

Let us compute the scalar product between the gradient and the direction \mathbf{d}_1 :

$$\frac{\partial f}{\partial \mathbf{d}_1} = \left\langle \begin{pmatrix} 2xe^y \\ x^2 e^y \end{pmatrix}, \begin{pmatrix} 0 \\ 1 \end{pmatrix} \right\rangle = \begin{pmatrix} 2xe^y \cdot 0 \\ x^2 e^y \cdot 1 \end{pmatrix} = \begin{pmatrix} 0 \\ x^2 e^y \end{pmatrix}$$

The intuition of this example is formalized in the following

Fact 20.0.1. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$. The directional derivative along a certain direction $\mathbf{d} \in \mathbb{R}^n$ can be computed as the scalar product between the gradient of the function and the direction, formally

$$\frac{\partial f}{\partial \mathbf{d}} = \langle \nabla f, \mathbf{d} \rangle$$

Definition 20.0.1 (Vector-valued function). A function which codomain is multi-dimensional is called **vector-valued function**. Formally, $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$,

$$\text{where } f(\mathbf{x}) = \begin{pmatrix} f_1(\mathbf{x}) \\ f_2(\mathbf{x}) \\ \vdots \\ f_m(\mathbf{x}) \end{pmatrix} \in \mathbb{R}^m.$$

For such functions the computation of the derivative requires to specify not only the component with respect to the one the derivation should be performed, but also the index of the function.

Definition 20.0.2 (Partial derivative for vector-valued functions). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$, the partial derivative of the j -th function with respect to the i -th component is

$$\frac{\partial f_j}{\partial x_i}(\mathbf{x}) = \lim_{t \rightarrow 0} \frac{f_j(x_1, x_2, \dots, x_{i-1}, \dots, x_i + t, x_{i+1}, \dots, x_n) - f_j(\mathbf{x})}{t}$$

where $t \in \mathbb{R}$.

Definition 20.0.3 (Jacobian). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$ be a vector-valued function. We call **Jacobian** the matrix of all its first-order partial derivatives.

$$Jf(\mathbf{x}) = \begin{pmatrix} \frac{\partial f_1}{\partial x_1} & \dots & \frac{\partial f_1}{\partial x_n} \\ \vdots & \ddots & \vdots \\ \frac{\partial f_m}{\partial x_1} & \dots & \frac{\partial f_m}{\partial x_n} \end{pmatrix} = \begin{pmatrix} \nabla f_1(\mathbf{x}) \\ \nabla f_2(\mathbf{x}) \\ \vdots \\ \nabla f_m(\mathbf{x}) \end{pmatrix}$$

Example 20.0.2. Let $f : \mathbb{R}^2 \rightarrow \mathbb{R}$ such that $f(x, y) = x^2 e^y$ as before, where the gradient was

$$\nabla f(\mathbf{x}) = \begin{pmatrix} \frac{\partial f}{\partial x} \\ \frac{\partial f}{\partial y} \end{pmatrix} = \begin{pmatrix} 2xe^y \\ x^2 e^y \end{pmatrix}$$

Let us compute the second derivative of this function:

$$\nabla^2 f(\mathbf{x}) = \begin{pmatrix} \frac{\partial^2 f}{\partial x \partial x} & \frac{\partial^2 f}{\partial y \partial x} \\ \frac{\partial^2 f}{\partial x \partial y} & \frac{\partial^2 f}{\partial y \partial y} \end{pmatrix} = \begin{pmatrix} 2e^y & 2xe^y \\ 2xe^y & x^2 e^y \end{pmatrix}$$

The second order derivative of a vector function is a matrix, defined as

Definition 20.0.4 (Hessian). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$, we call **Hessian** of f

$$\nabla^2 f(\mathbf{x}) := J \nabla f(\mathbf{x}) = \begin{pmatrix} \frac{\partial^2 f}{\partial x_1^2}(\mathbf{x}) & \frac{\partial^2 f}{\partial x_2 \partial x_1}(\mathbf{x}) & \dots & \frac{\partial^2 f}{\partial x_n \partial x_1}(\mathbf{x}) \\ \frac{\partial^2 f}{\partial x_1 \partial x_2}(\mathbf{x}) & \frac{\partial^2 f}{\partial x_2^2}(\mathbf{x}) & \dots & \frac{\partial^2 f}{\partial x_n \partial x_2}(\mathbf{x}) \\ \vdots & \vdots & \ddots & \vdots \\ \frac{\partial^2 f}{\partial x_1 \partial x_n}(\mathbf{x}) & \frac{\partial^2 f}{\partial x_2 \partial x_n}(\mathbf{x}) & \dots & \frac{\partial^2 f}{\partial x_n^2}(\mathbf{x}) \end{pmatrix}$$

The size of the matrix grows very rapidly with the number of derivatives that we make.

In optimization, handling Hessians is a crucial task, because such matrices are very large hence unfeasible most of the times.

In most cases, the Hessian is symmetric matrix, as stated in the following

Theorem 20.0.2 (Schwartz's theorem). *Let $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$ such that $\nabla^2 f(\mathbf{x})$ exists and it is continuous and let us take $\mathbf{x} \in \mathbb{R}^n$. $\exists \delta > 0$ s.t. $\forall \mathbf{x}' \in \mathcal{B}(\mathbf{x}, \delta)$ and $\forall i, j \in \{1, 2, \dots, n\}$ $\frac{\partial^2 f}{\partial x_j \partial x_i}(\mathbf{x}')$ and $\frac{\partial^2 f}{\partial x_i \partial x_j}(\mathbf{x}')$ exist and they are equal:*

$$\frac{\partial^2 f}{\partial x_j \partial x_i}(\mathbf{x}') = \frac{\partial^2 f}{\partial x_i \partial x_j}(\mathbf{x}')$$

Definition 20.0.5 (\mathcal{C}^2 functions). *Let $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$. We say that f belongs to \mathcal{C}^2 class iff $\nabla^2 f(x)$ is continuous.*

Corollary 20.0.3. *Let $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$ such that $f \in \mathcal{C}^2$, then the Hessian is symmetric and the gradient is continuous.*

We are now interested in providing some approximations to a generic function f . Such approximations are useful tools to have a hint on where the function moves along a decreasing direction, but the information is accurate only locally.

Definition 20.0.6 (First order Taylor model). *Let $f : \mathbb{R}^n \rightarrow \mathbb{R} \in \mathcal{C}^1$, we define the **first-order Taylor's formula** as a linear approximation of the function f in a neighbourhood of \mathbf{x} .*

Formally, for a given ball $\mathcal{B}(\mathbf{x}, \delta)$, $\forall \mathbf{x}' \in \mathcal{B}(\mathbf{x}, \delta) \exists \alpha \in (0, 1)$ s.t.

$$f(\mathbf{x}') \approx L_{\mathbf{x}}(\mathbf{x}') = \langle \nabla f(\alpha \mathbf{x} + (1 - \alpha)\mathbf{x}'), \mathbf{x}' - \mathbf{x} \rangle + f(\mathbf{x})$$

*Equivalently, we can write the so-called **remainder version of first-order Taylor formula** as*

$$f(\mathbf{x}') \approx L_{\mathbf{x}}(\mathbf{x}') = \langle \nabla f(\mathbf{x}), \mathbf{x}' - \mathbf{x} \rangle + f(\mathbf{x}) + R_2(\mathbf{x}') \quad (20.0.1)$$

*where R_2 is called **quadratic remainder** and is such that $\lim_{h \rightarrow 0} \frac{R(h)}{\|h\|} = 0$.*

Notice the general Lagrangian form of the remainder:

$$R_k(\mathbf{x}') = \frac{f^{(k+1)}(\beta)}{(k+1)!} \cdot (\mathbf{x}' - \mathbf{x})^2$$

where $\beta \in [\mathbf{x}, \mathbf{x}']$ and f is $k+1$ times differentiable.

Notice that the Taylor's approximation works *only locally*: the furthest we get from \mathbf{x} the more distant the function and the model are.

Definition 20.0.7 (Second order Taylor model). *Let $f : \mathbb{R}^n \rightarrow \mathbb{R} \in \mathcal{C}^1$, we define the **second-order Taylor's formula** as a quadratic approximation of the function f in a neighbourhood of \mathbf{x} . Formally, given a ball $\mathcal{B}(\mathbf{x}, \delta)$, $\forall \mathbf{x}' \in \mathcal{B}(\mathbf{x}, \delta)$*

$$f(\mathbf{x}') \approx Q_{\mathbf{x}}(\mathbf{x}') = L_{\mathbf{x}}(\mathbf{x}') + \frac{1}{2}(\mathbf{x}' - \mathbf{x})^T \nabla^2 f(\mathbf{x})(\mathbf{x}' - \mathbf{x}) + R_3(\mathbf{x}' - \mathbf{x}) \quad (20.0.2)$$

with $\lim_{h \rightarrow 0} R(h) \|h\|^2 = 0$ or, equivalently, the remainder vanishes at least cubically, or the error is $O(\|\mathbf{x}' - \mathbf{x}\|^3)$.

Example 20.0.3. Let $f : \mathbb{R}^2 \rightarrow \mathbb{R}$ such that $f(\mathbf{x}) = f\begin{pmatrix} x_1 \\ x_2 \end{pmatrix} = x_1^2 e^{x_2}$, as above and $\nabla f(\mathbf{x}) = \begin{pmatrix} 2x_1 e^{x_2} \\ x_1^2 e^{x_2} \end{pmatrix}$ and $\nabla^2 f(\mathbf{x}) = \begin{pmatrix} 2e^{x_2} & 2x_1 e^{x_2} \\ 2x_1 e^{x_2} & x_1^2 e^{x_2} \end{pmatrix}$. In $\mathbf{x} = \begin{pmatrix} 1 \\ 0 \end{pmatrix} \in \mathbb{R}^2$ we have $f(\mathbf{x}) = 1$ and $\nabla f(\mathbf{x}) = \begin{pmatrix} 2 \\ 1 \end{pmatrix}$.

Compute the first and second order Taylor models for f with $\alpha = 1$.

The linear model has the following shape

$$L_{(1,0)}(x_1, x_2) = \langle \begin{pmatrix} 2 \\ 1 \end{pmatrix}, \begin{pmatrix} x_1 - 1 \\ x_2 - 0 \end{pmatrix} \rangle + 1 = 2x_1 - 2 + x_2 + 1 = 2x_1 + x_2 - 1$$

And the quadratic model is

$$f(\mathbf{x}') = \langle \nabla f\left(\alpha \cdot \begin{pmatrix} 2 \\ 1 \end{pmatrix} + (1-\alpha)\mathbf{x}'\right), \mathbf{x}' - \begin{pmatrix} 2 \\ 1 \end{pmatrix} \rangle + f\left(\begin{pmatrix} 2 \\ 1 \end{pmatrix}\right) + \frac{1}{2} \cdot \nabla^2 f\left(\begin{pmatrix} 2 \\ 1 \end{pmatrix}\right) \cdot \left(\mathbf{x}' - \begin{pmatrix} 2 \\ 1 \end{pmatrix}\right)$$

$$\begin{aligned} Q_{[1,0]}(x_1, x_2) &= 2x_1 + x_2 - 1 + \frac{1}{2} \cdot (x_1 - 1, x_2 - 0) \cdot \begin{pmatrix} 2 & 2 \\ 2 & 1 \end{pmatrix} \cdot \begin{pmatrix} x_1 - 1 \\ x_2 - 0 \end{pmatrix} \\ &= 2x_1 + x_2 - 1 + \frac{1}{2} \cdot (2x_1 - 2 + 2x_2, 2x_1 - 2 + x_2) \cdot \begin{pmatrix} x - 1 \\ x_2 - 0 \end{pmatrix} \\ &= 2x_1 + x_2 - 1 + \frac{1}{2} \cdot ((2x_1 - 2 + 2x_2) \cdot (x_1 - 1) + (2x_1 - 2 + x_2) \cdot x_2) \\ &= 2x_1 + x_2 - 1 + \frac{1}{2} \cdot (2x_1^2 - 2x_1 + 2x_1x_2 - 2x_1 + 2 - 2x_2 + 2x_1x_2 - 2x_2 + x_2^2) \\ &= 2x_1 + x_2 - 1 + \frac{1}{2} \cdot (2x_1^2 - 4x_1 + 4x_1x_2 + 2 - 4x_2 + x_2^2) \\ &= 2x_1 + x_2 - 1 + x_1^2 - 2x_1 + 2x_1x_2 + 1 - 2x_2 + \frac{1}{2}x_2^2 \\ &= x_1^2 + 2x_1x_2 - x_2 + \frac{1}{2}x_2^2 \end{aligned} \tag{20.0.3}$$

The Taylor model can be extended to the more general k -th order. Such an expansion requires the computation of the k -th order derivatives, but $\nabla^k f(\mathbf{x})$ is a tensor of order $k \equiv n^k$ numbers. For $k > 2$ this approach is practically unfeasible.

As it happens often in computer science, it is important to find a good trade-off between the complexity of the model and the accuracy of the approximation: if the model is too simple (but efficient) the approximation is not very good; conversely, complex and accurate approximations are computationally heavy.

Fact 20.0.4. Let $f : \mathbb{R}^n \rightarrow \mathbb{R} \in \mathcal{C}^1$. If f is Lipschitz continuous on $S \in \mathbb{R}^n$, then in such set the norm of the gradient is bounded by the Lipschitz constant.

Formally, $\sup\{\|\nabla f(\mathbf{x})\| : \mathbf{x} \in S\} \leq L$. If S convex $\sup\{\|\nabla f(\mathbf{x})\| : \mathbf{x} \in S\} = L$.

Moreover, we can prove

Fact 20.0.5. Let $f : \mathbb{R}^n \rightarrow \mathbb{R} \in \mathcal{C}^1$. f is Lipschitz continuous on $S \in \mathbb{R}^n$ iff its Hessian is bounded on S . Formally, $\sup\{\|\nabla^2 f(\mathbf{x})\| : \mathbf{x} \in S\} \leq L$.

Fact 20.0.6. Let $f : \mathbb{R}^n \rightarrow \mathbb{R} \in \mathcal{C}^1$. If the gradient of f is Lipschitz continuous in all points $\mathbf{x}' \in \mathbb{R}^n$ that are close to $\mathbf{x} \in \mathbb{R}^n$, then $f(\mathbf{x}') \leq L_x(\mathbf{x}') + \frac{L}{2} \|\mathbf{x}' - \mathbf{x}\|^2$.

20.1 Simple functions

In the rest of the course we will deal mostly with some linear and quadratic approximations of the objective function. In this section, we introduce a couple of examples and considerations on such functions.

20.1.1 Linear functions

In this scenario, $f : \mathbb{R}^n \rightarrow \mathbb{R}$ has the following shape: $f(\mathbf{x}) = \mathbf{c}^T \mathbf{x}$, for a fixed $\mathbf{c} \in \mathbb{R}^n$.

It holds that $\nabla f(\mathbf{x}) = \mathbf{c}$, $\nabla^2 f(\mathbf{x}) = 0$ and that level sets are parallel hyperplanes orthogonal to \mathbf{c} .

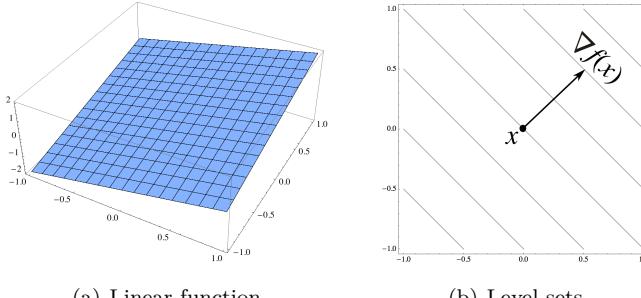


FIGURE 20.1: Graphical example of linear function.

20.1.2 Quadratic functions

A quadratic function $f : \mathbb{R}^n \rightarrow \mathbb{R}$ is formalized as $f(\mathbf{x}) = \frac{1}{2} \mathbf{x}^T Q \mathbf{x} + \mathbf{q}^T \mathbf{x}$, where $Q \in \mathcal{M}(n, \mathbb{R})$, $\mathbf{q}^T \in \mathbb{R}^n$.

In Figure 20.2 we can see the plot of the quadratic function where

$$Q = \begin{bmatrix} 6 & -2 \\ -2 & 6 \end{bmatrix}, \mathbf{q} = \begin{bmatrix} 0 \\ 0 \end{bmatrix}$$

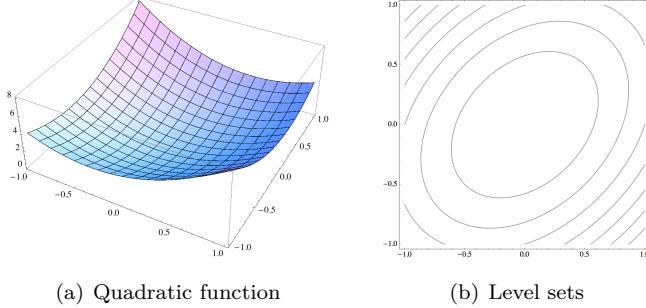


FIGURE 20.2: Graphical example of a quadratic function.

In quadratic functions the gradient is a linear function on \mathbf{x} $\nabla f(\mathbf{x}) = Q\mathbf{x} + \mathbf{q}$, while the Hessian is just Q and the level sets are ellipsoids. Sometimes such ellipses can degenerate to lines, in the case that one of the axis has become $+\infty$.

Fact 20.1.1. *Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ be a quadratic function with parameters $Q \in \mathcal{M}(n, \mathbb{R})$ and $\mathbf{q}^T \in \mathbb{R}^n$. If Q is symmetric, then it has spectral decomposition.*

$$\text{Proof. } \mathbf{x}^T Q \mathbf{x} = \frac{[(\mathbf{x}^T Q \mathbf{x}) + (\mathbf{x}^T Q \mathbf{x})^T]}{2} = \mathbf{x}^T \left[\frac{(Q + Q^T)}{2} \right] \mathbf{x} = \mathbf{x}^T H \Lambda H^T$$

□

In the rest of the course we will assume for simplicity that Q is symmetric. Another “lucky case” is when Q is non singular (i.e. all its eigenvalues are not 0): in this case, $\bar{\mathbf{x}} = -Q^{-1}\mathbf{q}$, where $\bar{\mathbf{x}}$ is called the center of the ellipsoid. Let us assume that we moved the origin in such $\bar{\mathbf{x}}$, so $\mathbf{x}' = \mathbf{x} - \bar{\mathbf{x}}$ and $f_{\bar{\mathbf{x}}}(\mathbf{x}') = \frac{1}{2}\mathbf{x}'^T Q \mathbf{x}'$.

Fact 20.1.2. *Along H_i : $f(\alpha) = f_{\bar{\mathbf{x}}}(\alpha H_i) = \alpha^2 \lambda_i$*

Moreover, the size of the axes of the level curves is proportional to $\sqrt{1/\lambda_i}$, where λ_i are the eigenvalues. If the eigenvalues are close to 0, then the axes get very long, while when they are negative, we no longer have axes.

This is formalized in

Fact 20.1.3. *Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ such that $f(\mathbf{x}) = \frac{1}{2}\mathbf{x}^T Q \mathbf{x} + \mathbf{q}^T \mathbf{x}$. If Q is positive definite $\bar{\mathbf{x}} = -Q^{-1}\mathbf{q}$ is the minimum of f , while if Q is indefinite (i.e. $\exists \lambda_i < 0$) f is unbounded below.*

Lecture 21

5th of October 2018

A. Frangioni

21.1 Unconstrained optimization

Until now we stated that the best conditions for finding the optimum are encountered when the domain is a compact set and we have many derivatives.

Now we need to consider when we can stop our algorithm.

Definition 21.1.1 (Unconstrained optimization problem). *We term **unconstrained optimization problem** the following*

$$(P) f_* = \min\{f(\mathbf{x}) : \mathbf{x} \in \mathbb{R}^n\}$$

In unconstrained optimization, we deal with an unbounded set (\mathbb{R}^n), hence Weierstrass theorem does not apply. Because of this reason, we have no guarantee that a minimum \mathbf{x}_* exists; moreover, provided its existence, finding it is an NP-hard problem. In order to make things easier, in practice we use a weaker condition: **local minimality**.

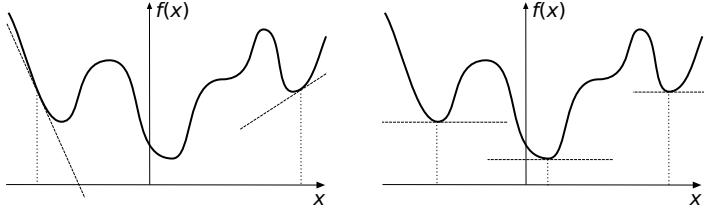
Definition 21.1.2 (Local minimum). *Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$. \mathbf{x}_* is a **local minimum** if it is a global minimum in a ball around \mathbf{x}^* . Formally,*

$$\min\{f(\mathbf{x}) : \mathbf{x} \in \mathcal{B}(\mathbf{x}_*, \varepsilon)\}$$

for some $\varepsilon > 0$.

Also, \mathbf{x}^* is a **strict local minimum** if $f(\mathbf{x}) < f(\mathbf{x}') \forall \mathbf{x}' \in \mathcal{B}(\mathbf{x}_*, \varepsilon)$.

If $f'(\mathbf{x}) < 0$ or $f'(\mathbf{x}) > 0$, \mathbf{x} clearly cannot be a local minimum, as shown in Figure 21.1.



(a) Points where the derivative is not 0.

(b) Points where the derivative is 0.

FIGURE 21.1: Minima and not minima.

21.1.1 First order model

💡 Do you recall?

The first order model of f is $L_{\mathbf{x}}(\mathbf{x}') = f(\mathbf{x}) + \nabla f(\mathbf{x})(\mathbf{x}' - \mathbf{x})$, such that $\forall \mathbf{x}' \in \mathbb{R}^n$ close to \mathbf{x} $f(\mathbf{x}') = L_{\mathbf{x}}(\mathbf{x}') + R(\mathbf{x}' - \mathbf{x})$, where $R(\cdot)$ is called **residual** and it has the property of quadratic convergence: $\lim_{\|\mathbf{h}\| \rightarrow 0} \frac{R(\mathbf{h})}{\|\mathbf{h}\|} = 0$.

Fact 21.1.1. Let f be differentiable, if \mathbf{x} is a local minimum, then $\nabla f(\mathbf{x}) = 0$.

In optimization, we are interested in moving towards a (local) minimum \mathbf{x}_* as fast as possible.

Fact 21.1.2. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ be the objective function of the optimization problem (P) and let $\mathbf{x} \in \mathbb{R}^n$ be the current point at a generic iteration. In order to get closer to the optimum, we need to take a step along the anti-gradient direction. Formally, $\mathbf{x}(\alpha) = \mathbf{x} - \alpha \nabla f(\mathbf{x})$, where $\alpha \in \mathbb{R}$ is called step size.

Proof by contradiction. Let us assume that \mathbf{x} is a local minimum but $\nabla f(\mathbf{x}) \neq 0$.

In our case, $\mathbf{x}' = \mathbf{x} - \alpha \nabla f(\mathbf{x})$, so let us plug it into the *remainder version* of the Tailor's first-order model:

$$\begin{aligned} f(\mathbf{x}') &= \langle \nabla f(\mathbf{x}), \mathbf{x}' - \mathbf{x} \rangle + f(\mathbf{x}) + R(\mathbf{x}' - \mathbf{x}) \\ &= \langle \nabla f(\mathbf{x}), \mathbf{x} - \alpha \nabla f(\mathbf{x}) - \mathbf{x} \rangle + f(\mathbf{x}) + R(\mathbf{x} - \alpha \nabla f(\mathbf{x}) - \mathbf{x}) \\ &= \langle \nabla f(\mathbf{x}), -\alpha \nabla f(\mathbf{x}) \rangle + f(\mathbf{x}) + R(-\alpha \nabla f(\mathbf{x})) \\ &= f(\mathbf{x}) - \alpha \|\nabla f(\mathbf{x})\|^2 + R(-\alpha \nabla f(\mathbf{x})) \end{aligned}$$

Once we fixed the moving direction, we can choose the step size α , so it can be proved that $\lim_{\alpha \rightarrow 0} \frac{R(-\alpha \nabla f(\mathbf{x}))}{\|\alpha \nabla f(\mathbf{x})\|} = 0$, that is equivalent by definition to $\forall \varepsilon > 0 \exists \bar{\alpha} > 0$ s.t. $\frac{R(-\alpha \nabla f(\mathbf{x}))}{\alpha \|\nabla f(\mathbf{x})\|} \leq \varepsilon \quad \forall 0 \leq \alpha < \bar{\alpha}$.

If we take $\varepsilon < \|\nabla f(\mathbf{x})\|$, we get $R(-\alpha \nabla f(\mathbf{x})) < \alpha \|\nabla f(\mathbf{x})\|^2$, then

$$f(\mathbf{x}(\alpha)) = f(\mathbf{x}) - \alpha \|\nabla f(\mathbf{x})\|^2 + R(-\alpha \nabla f(\mathbf{x})) < f(\mathbf{x})$$

$\forall \alpha < \bar{\alpha}$ \mathbf{x} cannot be a local minimum. \square

Proposition 21.1.2 states that the first order model allows to find the decreasing direction, but if the gradient is 0 we do not know if we are in presence of a minimum, maximum or a saddle point. To discriminate among those, we exploit the information provided by the second derivative.

21.1.2 Second order model

Fact 21.1.3. *Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ such that $f \in C^2$. If \mathbf{x} is a local minimum then the gradient is positive semidefinite. Formally, $\nabla^2 f(\mathbf{x}) \succeq 0$.*

Proof by contradiction. Our contradictory hypothesis is that we are in a local minimum, but the Hessian is not positive semidefinite, there there is a direction $\mathbf{d} \in \mathbb{R}^n$ of negative curvature. Formally, $\exists \mathbf{d}$ s.t. $\mathbf{d}^T \nabla^2 f(\mathbf{x}) \mathbf{d} < 0$ or equivalently, the Hessian has a negative eigenvalue $\lambda_i < 0$.

Let us consider that \mathbf{d} has norm 1 without loss of generality and let us be given a step $\mathbf{x}(\alpha) = \mathbf{x} + \alpha \mathbf{d}$ and then take the second-order Taylor formula as follows (where there is no linear term involved because $\nabla f(\mathbf{x}) = 0$):

$$\begin{aligned} f(\mathbf{x}(\alpha)) &= f(\mathbf{x}) + \frac{1}{2} (\alpha \mathbf{d})^T \nabla^2 f(\mathbf{x}) (\alpha \mathbf{d}) + R(\mathbf{x}(\alpha) - \mathbf{x}) \\ f(\mathbf{x}(\alpha)) &= f(\mathbf{x}) + \frac{1}{2} \alpha^2 \mathbf{d}^T \nabla^2 f(\mathbf{x}) \mathbf{d} + R(\alpha \mathbf{d}) \end{aligned}$$

with $\lim_{\|h\| \rightarrow 0} \frac{R(h)}{\|h\|^2} = \lim_{\alpha \rightarrow 0} \frac{R(\alpha \mathbf{d})}{\alpha^2} = 0$ or equivalently $\forall \varepsilon > 0 \exists \bar{\alpha} > 0$ s.t. $R(\alpha \mathbf{d}) \leq \varepsilon \alpha^2 \forall 0 \leq \alpha < \bar{\alpha}$.

At this point, since this condition holds for each ε we are allowed to take the most convenient: $\varepsilon < -\frac{1}{2} \mathbf{d}^T \nabla^2 f(\mathbf{x}) \mathbf{d}$, so that we obtain this condition on the residual $R(\alpha \mathbf{d}) < -\frac{1}{2} \alpha^2 \mathbf{d}^T \nabla^2 f(\mathbf{x}) \mathbf{d}$, hence

$$f(\mathbf{x}(\alpha)) = f(\mathbf{x}) + \frac{1}{2} \alpha^2 \mathbf{d}^T \nabla^2 f(\mathbf{x}) \mathbf{d} + R(\alpha \mathbf{d}) < f(\mathbf{x}) \forall 0 \leq \alpha < \bar{\alpha}$$

The contradiction lies in the fact that $\mathbf{x}(\alpha)$ leads to a value of f smaller than the one in \mathbf{x} . \square

Proposition 21.1.3 states that a positive semidefinite Hessian is a necessary condition for a local minimum. Follows a proposition that gives a sufficient condition.

Fact 21.1.4. *Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ s.t. $f \in C^2$ and let the Hessian be symmetric (hence real eigenvalues). If $\nabla f(\mathbf{x}) = 0$ and the Hessian is strictly positive definite ($\nabla^2 f(\mathbf{x}) \succ 0$) then \mathbf{x} is a local minimum.*

Proof. $\mathbf{x} \in \mathbb{R}^n$ is a point where the gradient of the function is 0 we get the following second order Taylor approximation for $\mathbf{x}_{\text{new}} = \mathbf{x} + \mathbf{d}$ (unitary step size for simplicity):

$$f(\mathbf{x} + \mathbf{d}) = f(\mathbf{x}) + \frac{1}{2}\mathbf{d}^T \nabla^2 f(\mathbf{x})\mathbf{d} + R(\mathbf{d}) \text{ with } \lim_{h \rightarrow 0} \frac{R(\mathbf{d})}{\|\mathbf{d}\|^2} = 0$$

Hence, by definition of limit $\forall \varepsilon > 0 \exists \delta > 0$ s.t. $R(\mathbf{d}) \leq \varepsilon \|\mathbf{d}\|^2 \forall \mathbf{d} \in \mathbb{R}^n$ s.t. $\|\mathbf{d}\| < \delta$.

Since the Hessian is strictly positive definite, the smallest eigenvalue λ_{\min} is strictly greater than 0, hence the variational characterization of eigenvalues $\mathbf{d}^T \nabla^2 f(\mathbf{x})\mathbf{d} \geq \lambda_{\min} \|\mathbf{d}\|^2$.

Aggiungere la ref alla variational characterization nella parte di Numerical Methods

We are now ready to pick the ε we prefer ($\varepsilon < \lambda_{\min}$) to get $\forall \mathbf{d}$ s.t. $\|\mathbf{d}\| < \delta$

$$f(\mathbf{x} + \mathbf{d}) = f(\mathbf{x}) + \frac{1}{2}\mathbf{d}^T \nabla^2 f(\mathbf{x})\mathbf{d} + R(\mathbf{d}) \geq f(\mathbf{x}) + (\lambda_{\min} - \varepsilon)\|\mathbf{d}\|^2 > f(\mathbf{x})$$

The term $\lambda_{\min} - \varepsilon$ is strictly positive. □

In the rest of this lecture, we will provide conditions that ensure that once a local minimum is found, it is also a global minimum.

So far, we said that the local minima are those points where the gradient is 0 and the Hessian is positive semidefinite. An easy way to ensure that the Hessian is positive semidefinite in a ball around \mathbf{x} is to have that the Hessian is positive semidefinite everywhere ($\forall \mathbf{x} \in \mathbb{R}^n$) and this is guaranteed to be true whenever f is a convex function.

21.2 Convexity

Let us introduce convexity as far as both sets and functions are concerned.

21.2.1 Convex sets

Definition 21.2.1 (Convex hull). *Let $\mathbf{x}, \mathbf{y} \in \mathbb{R}^n$ we term **convex hull** and denote $\text{conv}(\mathbf{x}, \mathbf{y}) = \{\mathbf{z} = \alpha\mathbf{x} + (1 - \alpha)\mathbf{y} : \alpha \in [0, 1]\}$ the segment joining \mathbf{x} and \mathbf{y} .*

Definition 21.2.2 (Convex set). *Intuitively, we say that C is a **convex set** if for each couple in the set, the line joining such points belongs to the set. Formally, $C \subset \mathbb{R}^n$ is a **convex set** if $\forall \mathbf{x}, \mathbf{y} \in C \text{ } \text{conv}(\mathbf{x}, \mathbf{y}) \subseteq C$.*

Notice that “disconnected sets” cannot be convex sets.

Definition 21.2.3 (Convex hull of a set). *Given a set $S \subseteq \mathbb{R}^n$, we can “complete” it to a convex set $\text{conv}(S)$, called **convex hull of S** :*

$$\begin{aligned}\text{conv}(S) &= \bigcup\{\text{conv}(\mathbf{x}, \mathbf{y}) : \mathbf{x}, \mathbf{y} \in S\} \\ &= \bigcap\{C : C \text{ is convex} \wedge C \supseteq S\}\end{aligned}$$

Equivalently, the convex hull of S can be defined iteratively as the convex hull of all the couples of points in S .

Equivalently, we can define the convex hull of a set S as the smallest convex set containing S .

Note

A more general definition of a convex hull is the following:

$$\text{conv}(\{\mathbf{x}_1, \dots, \mathbf{x}_k\}) = \left\{ \mathbf{x} = \sum_{i=1}^k \alpha_i \mathbf{x}_i : \sum_{i=1}^k \alpha_i = 1, \alpha_i \geq 0 \forall i \right\}$$

Definition 21.2.4 (Unitary simplex). *We term **unitary simplex** the set of k non-negative scalars summing to 1, formally*

$$\Theta^k = \left\{ \sum_{i=1}^k \alpha_i \mathbf{x}_i \in \mathbb{R}^k : \sum_{i=1}^k \alpha_i = 1, \alpha_i \geq 0 \forall i \right\}$$

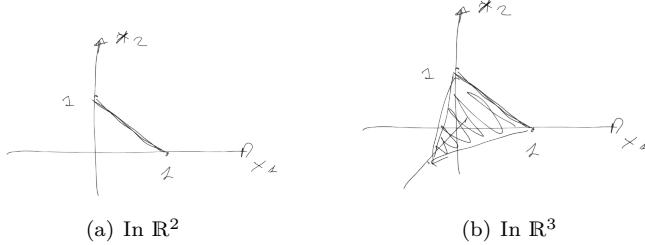


FIGURE 21.2: Example of unitary simplexes.

Our goal is to find the simplest possible convex set that approximates our set.

Fact 21.2.1. *A convex set is equal to its convex hull. Formally,*

$$C \text{ is convex} \iff C = \text{conv}(C)$$

We are interested in sufficient conditions for convexity.

Definition 21.2.5 (Cone). *We term **cone** the set $\mathcal{C} = \{\mathbf{x} \in \mathbb{R}^n : \alpha \mathbf{x} \in \mathcal{C} \forall \alpha \geq 0\}$.*

An attentive reader may notice that the definition of cone is a relaxation of the unitary simplex, where we do not require the unitary sum.

Example 21.2.1. *The following sets are convex:*

- Convex polytope $\text{conv}(\{\mathbf{x}_1, \dots, \mathbf{x}_k\})$, unitary simplex Θ
- Affine hyperplane: $\mathcal{H} := \{\mathbf{x} \in \mathbb{R}^n : \mathbf{vect}\mathbf{x} = b\}$
- Affine close (open) subspace: $\mathcal{S} := \{\mathbf{x} \in \mathbb{R}^n : \mathbf{a}^T \mathbf{vect}\mathbf{x} \leq (<)b\}$
- Close (open) ball of radius r in p -norm, $p \geq 1$: $\mathcal{B}_p(\mathbf{x}, r) = \{\mathbf{y} \in \mathbb{R}^n : \|\mathbf{y} - \mathbf{x}\|_p \leq (<)r\}$
- Close (open) ellipsoid: $\mathcal{E}(Q, \mathbf{x}, r) := \{y \in \mathbb{R}^n : (y - \mathbf{x})^T Q (y - \mathbf{x}) \leq (<)r\}$ with $Q \succeq 0$. Notice that ellipsoids are levelsets of quadratic functions.
- Cones
- Conical hull of a finite set of directions:

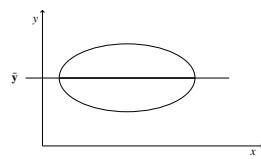
$$\text{cone}(\{\mathbf{d}_1, \dots, \mathbf{d}_k\}) = \left\{ \mathbf{d} = \sum_{i=1}^k \mu_i \mathbf{d}_i : \mu_i \geq 0 \forall i \right\}$$

- Lorentz's cone (ice-cream cone): $\mathbb{L} = \left\{ \mathbf{x} \in \mathbb{R}^n : \mathbf{x}_n \geq \sqrt{\sum_{i=1}^{n-1} \mathbf{x}_i^2} \right\}$
- Cone of positive semidefinite matrices: $\mathbb{S}_+ = \{A \in (M)(n, \mathbb{R}^n) : A \succeq 0\}$

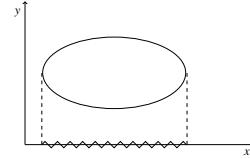
Fact 21.2.2. *The following operations preserve convexity.*

1. Given a possibly infinite family of convex sets $(\{C_i\}_{i \in I})$, the intersection $(\bigcap_{i \in I} C_i)$ is convex;
2. If we have convex sets in different subspaces, their Cartesian product is a convex set (C_1, \dots, C_k convex $\iff C_1 \times \dots \times C_k$ convex);
3. Given a convex set, its image under a linear mapping (i.e. scaling, translation, rotation) is a convex set. Formally, let C be a convex set. Then $A(C) := \{\mathbf{x} = A\mathbf{y} + \mathbf{b} : \mathbf{y} \in C\}$ for a given $A \in M(n, \mathbb{R})$ and $\mathbf{b} \in \mathbb{R}^n$ is convex;
4. If C is a convex set its inverse image under a linear mapping is convex as well. Formally, $A^{-1}(C) := \{x : Ax + b \in C\}$ is convex;
5. Let C_1 and C_2 convex and let $\alpha_1, \alpha_2 \in \mathbb{R}$. Then $\alpha_1 C_1 + \alpha_2 C_2 := \{\mathbf{x} = \alpha_1 \mathbf{x}_1 + \alpha_2 \mathbf{x}_2 : \mathbf{x}_1 \in C_1, \mathbf{x}_2 \in C_2\}$ is convex;
6. Let $C \subseteq \mathbb{R}^{n_1} \times \mathbb{R}^{n_2}$ be a convex set. Slicing and projection lead to convex sets as well, as shown in Figure 21.3:

SLICE: $C_s(\mathbf{y}) := \{\mathbf{x} \in \mathbb{R}^{n_1} : (\mathbf{x}, \mathbf{y}) \in C\}$ is convex



(a) Slice on \bar{y}



(b) Project on the x component

FIGURE 21.3: Pictorial examples of slicing and projecting.

PROJECTION: $C_p := \{\mathbf{x} \in \mathbb{R}^{n_2} : \exists \mathbf{y} \text{ s.t. } (\mathbf{x}, \mathbf{y}) \in C\}$ is convex

7. Given a convex set C , the interior of such set and its closure are convex.

Theorem 21.2.3. \mathcal{P} is a polyhedron iff any of its points can be written as a combination of a point in a convex set and one in a cone. Formally, iff $\exists \{\mathbf{x}_1, \dots, \mathbf{x}_k\} \subseteq \mathbb{R}^n$ and $\{\mathbf{d}_1, \dots, \mathbf{d}_h\} \subseteq \mathbb{R}^n$ s.t. $\mathcal{P} = \text{conv}(\{\mathbf{x}_1, \dots, \mathbf{x}_k\}) + \text{cone}(\{\mathbf{d}_1, \dots, \mathbf{d}_h\})$.

Notice that whenever we are interested in proving that a set with a certain shape is convex, we could try to derive it from an object that we know is convex through the operations we enumerated above.

21.2.2 Convex functions

In the rest of this lecture we will introduce the class of convex functions, that have the characteristic of having local minima coinciding with global minima.

Definition 21.2.6 (Graph). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$, we term **graph** of f the set of ordered pairs (\mathbf{x}, y) such that $f(\mathbf{x}) = y$.

Definition 21.2.7 (Epigraph). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$, we term **epigraph** or **super-graph** of f the set of points lying on or above its graph. Formally,

$$\text{epi}(f) = \{(\mathbf{x}, \mu) : \mathbf{x} \in \mathbb{R}^n, \mu \in \mathbb{R}, \mu \geq f(\mathbf{x})\} \subseteq \mathbb{R}^{n+1}$$

Definition 21.2.8 (Convex function). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ be a function. We say that f is **convex** if $\forall \mathbf{x}, \mathbf{y} \in \mathbb{R}^n$, the segment that joins $f(\mathbf{x})$ and $f(\mathbf{y})$ lies above the function.

In other words, f is **convex** iff $\text{epi}(f)$ is convex.

Equivalently, we say that f is **convex** if $\forall \mathbf{x}, \mathbf{y} \in \text{dom}(f)$ for any $\alpha \in [0, 1]$, $\alpha f(\mathbf{x}) + (1 - \alpha)f(\mathbf{y}) \geq f(\alpha\mathbf{x} + (1 - \alpha)\mathbf{y})$.

Equivalently, $\forall \mathbf{x}^1, \dots, \mathbf{x}^k, \alpha \in \Theta^k$

$$f \left(\sum_{i=1}^k \alpha_i \mathbf{x}^i \right) \leq \sum_{i=1}^k \alpha_i f(\mathbf{x}^i)$$

Definition 21.2.9 (Strict convexity). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$. We term f **strictly convex** iff $\alpha f(\mathbf{x}) + (1 - \alpha)f(\mathbf{y}) > f(\alpha\mathbf{x} + (1 - \alpha)\mathbf{y})$.

Definition 21.2.10 (Concave function). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ be a function. We say that f is **concave** if $\forall \mathbf{x}, \mathbf{y} \in \mathbb{R}^n$, the segment that joins $f(\mathbf{x})$ and $f(\mathbf{y})$ lies below the function. Formally, $\forall \mathbf{x}, \mathbf{y} \in \text{dom}(f)$ for any $\alpha \in [0, 1]$, $\alpha f(\mathbf{x}) + (1 - \alpha)f(\mathbf{y}) \leq f(\alpha\mathbf{x} + (1 - \alpha)\mathbf{y})$.

Definition 21.2.11 (Sublevel graph). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$ be a function. We term **sublevel graph** of f on a scalar $v \in \mathbb{R}$ and denote $S(f, v)$ the projection on the \mathbf{x} axis of the portions of the graph of f which lie below the constant value v .

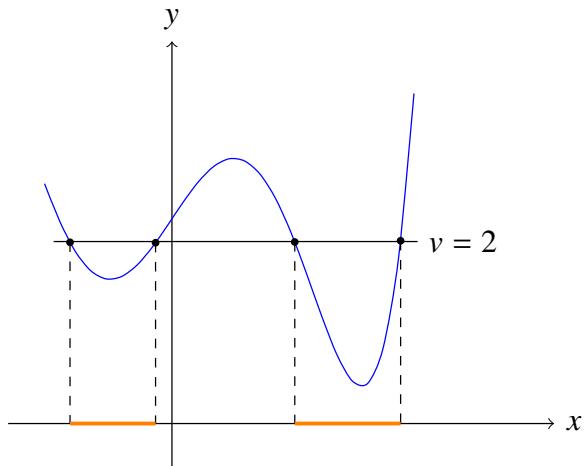


FIGURE 21.4: $f(x) = x^5 - 8x^3 + 10x + 15$ in blue and its sublevel graph in orange.

Fact 21.2.4. The following holds:

- Let f convex. Then $S(f, v)$ is convex $\forall v \in \mathbb{R}$;
- f is concave if $-f$ is convex.

★ Mantra

Convex analysis is a one-sided world.

The second statement of Proposition 21.2.4 is useful to make a comparison between minimizing and maximizing. In particular, if our aim is to maximize the function, we can be sure to have found a global maximum if the function is concave.

Definition 21.2.12 (Strong convexity). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$. We term f **strongly convex modulus** a scalar $\tau > 0$ iff $f(\mathbf{x}) - \frac{\tau}{2} \|\mathbf{x}\|^2$ is convex.

Formally,

$$\alpha f(\mathbf{x}) + (1 - \alpha)f(\mathbf{y}) \geq f(\alpha\mathbf{x} + (1 - \alpha)\mathbf{y}) + \frac{\tau}{2}\alpha(1 - \alpha)\|(\mathbf{y} - \mathbf{x})\|^2$$

In the next lecture we will provide techniques that allow to check if a function is convex in practice.

Lecture 22

11th of October 2018

A. Frangioni

In the previous lecture we introduced many different definitions about convexity, while in this chapter we will deal with the task of checking when a function is convex, hence a local minimum is also a global minimum.

As we already stated for convex sets, we can prove that a function is convex if we are able to derive it from convex functions, through “convex friendly” operations.

Note

A useful tool to have insights about convexity is the software called CVX, designed to model convex objects. A pretty easy way to check if an object is convex is to try to write it in CVX. If such an operation is possible, then the object is convex.

Let us enumerate some convex functions:

Linear functions: $f(\mathbf{x}) = \mathbf{w}^T \mathbf{x}$ (they are both convex and concave);

Quadratic functions: $f(\mathbf{x}) = \frac{1}{2} \mathbf{x}^T Q \mathbf{x} + \mathbf{q}^T \mathbf{x}$ is convex iff $Q \succeq 0$;

Exponential function: $f(\mathbf{x}) = e^{a\mathbf{x}}$ for any $a \in \mathbb{R}$ and $\mathbf{x} \in \mathbb{R}^n$

Anti-logarithmic function: $f(x) = -\log(x)$ for $x > 0$

Monomial: $f(x) = x^a$ for $a \geq 1$ or $a \leq 0$ on $x \geq 0$;

p -norm: $f(\mathbf{x}) = \|\mathbf{x}\|_p$ for $p \geq 1$;

Maximum: $f(x) = \max\{x_1, \dots, x_n\}$;

Indicator function: for any convex set C , its indicator function $\mathbb{1}_C(x)$ is convex:

$$\mathbb{1}_C(x) = \begin{cases} 0 & \text{if } x \in C \\ +\infty & \text{if } x \notin C \end{cases}$$

Sum of m largest eigenvalues: Let $A \in S(n, \mathbb{R})$ such that its eigenvalues (sorted in increasing order) are $\lambda_1 \geq \lambda_2 \geq \dots \lambda_n$. $f_m(A) = \sum_{i=1}^m \lambda_i$ is soncex.

Fact 22.0.1. *The following operations preserve convexity:*

Linear non-negative combination: let $f, g : \mathbb{R}^n \rightarrow \mathbb{R}$ be convex functions and let $\alpha, \beta \in \mathbb{R}_+$, then $\alpha f + \beta g$ is convex ;

Supremum: let $\{f_i\}_{i \in I}$ be a set of infinitely many convex functions, then $\sup_{i \in I} f_i(\mathbf{x})$ is convex, see Figure 22.1(a);

Pre-composition with linear function: let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ be a convex function, then $f(A\mathbf{x} + \mathbf{b})$ is convex;

Post-composition with increasing convex function: let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ be a convex function and let $g : \mathbb{R} \rightarrow \mathbb{R}$ be a convex increasing function. Then $g \circ f = g(f(\mathbf{x}))$ is convex;

Infimal convolution: let $f_1, f_2 : \mathbb{R}^n \rightarrow \mathbb{R}$ be convex functions. Then $f(\mathbf{x}) = \inf\{f_1(\mathbf{x}_1) + f_2(\mathbf{x}_2) : \mathbf{x}_1 + \mathbf{x}_2 = \mathbf{x}\}$ is convex;

Image under linear mapping: let $g : \mathbb{R}^n \rightarrow \mathbb{R}$ be a convex function. Then $f(\mathbf{x}) = \inf\{g(\mathbf{y}) : A\mathbf{y} = \mathbf{x}\}$ is convex;

Partial minimization: let $g \begin{pmatrix} \mathbf{x} \\ \mathbf{y} \end{pmatrix} : \mathbb{R}^{n+m} \rightarrow \mathbb{R}$ be a convex function. Then $f(\mathbf{x}) = \inf \{g \begin{pmatrix} \mathbf{x} \\ \mathbf{y} \end{pmatrix} : \mathbf{y} \in \mathbb{R}^m\}$ is convex;

Perspective or dilation: let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ be a convex function. Then

$$\tilde{f}(\mathbf{x}, u) = \begin{cases} u \cdot f(\mathbf{x}/u) & \text{if } u > 0 \\ \infty & \text{otherwise} \end{cases}$$

is convex, see Figure 22.1(b).

Fact 22.0.2. Let $f : \mathbb{R}^n \rightarrow \mathbb{R} \cup \infty$ be a convex function. If $\exists \bar{\mathbf{x}} \in \text{dom}(f)$ such that $f(\bar{\mathbf{x}}) = -\infty$, then $f \equiv -\infty$.

From now on we will solve the issue of functions with non a convex domain, saying that in those points where the function is not defined, we value it $+\infty$.

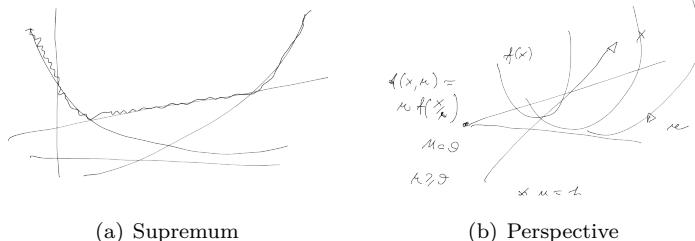
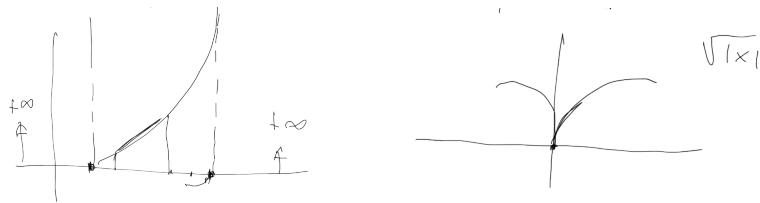


FIGURE 22.1: Examples of convexity-preserving operations.

Fact 22.0.3. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ be a convex function. Then f is Lipschitz continuous on each bounded convex set $S \subseteq \text{int}(\text{dom}(f))$, but we have no information on the behaviour on the border of the domain.

Moreover, a function f , which is continuous but not Lipschitz continuous is not convex.

A couple of examples of Proposition 22.0.3 can be found in Figure 22.2.



(a) On a compact set in the interior of the domain (far from the boundaries) the function is Lipschitz continuous.

(b) If a function is not Lipschitz on a compact subset it is not convex.

FIGURE 22.2

Fact 22.0.4. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ be a convex function. Then it is Lipschitz continuous on any bounded set and continuous everywhere.

It happens often that the set of points in which a function is non differentiable has measure 0.

Theorem 22.0.5 (Convexity characterization). *Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ be in \mathcal{C}^1 . It is convex on a convex set $C \subseteq \mathbb{R}^n$ iff*

$$f(\mathbf{y}) \geq f(\mathbf{x}) + \nabla f(\mathbf{x})(\mathbf{y} - \mathbf{x}) \quad \forall \mathbf{x}, \mathbf{y} \in C$$

Riscrivere proof

Proof. \Rightarrow) $\alpha(f(y) - f(x)) \geq f(\alpha(y - x) + x) - f(x)$, send $\alpha \rightarrow 0$

\Leftrightarrow) TODO

□

Theorem 22.0.6. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ in C^1 be a convex function. $\mathbf{x} \in \mathbb{R}^n$ is a stationary point for f iff \mathbf{x} is a global minimum.

Fact 22.0.7. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ be twice differentiable ($f \in C^2$). f is convex on an open set $S \subseteq \mathbb{R}^n$ iff the Hessian is positive semidefinite.

This proposition gives us an algorithm to check if a function is convex or not: we only need to compute the eigenvalues of the Hessian and check if they are positive.

There are some functions which do not have differentiability property.

A way to work with functions which are not defined on all \mathbb{R}^n is to solve the following problem:

$$(P) \equiv \inf\{f_X(\mathbf{x}) = f(\mathbf{x}) + \mathbf{1}_X(\mathbf{x}) : \mathbf{x} \in \mathbb{R}^n\}$$

thanks to the following

Theorem 22.0.8 (Essential objective). \mathbf{x}_* optimal for (P) $\iff \mathbf{x}_*$ local minimum of f_X .

22.0.1 Non-differentiable functions

It is often the case that the objective function under consideration is not differentiable and in such situation we need to use different tools to be sure to move towards the optimum. It is in this context that we introduce *subgradients* and *subdifferentials*.

Definition 22.0.1 (Subgradient). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$. We say that $\mathbf{s} \in \mathbb{R}^n$ is a *subgradient* of f at point $\mathbf{x} \in \mathbb{R}^n$ if $\forall \mathbf{y} \in \mathbb{R}^n$ the following holds:

$$f(\mathbf{y}) \geq f(\mathbf{x}) + \mathbf{s}(\mathbf{y} - \mathbf{x})$$

Let us assume that the minimum of the non differentiable function resides in one of its kinky points, then for $\mathbf{s} = \mathbf{0}$ we have a subgradient which is flat and this is a sufficient condition for minimality, for a pictorial example see Figure 22.3.

The issue here is that it is unfeasible to check if $\mathbf{s} = \mathbf{0}$ is a subgradient for f , since we should check all possible $\mathbf{y} \in \mathbb{R}^n$.

Definition 22.0.2 (Subdifferential). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ and let $\mathbf{x} \in \mathbb{R}^n$. We call *subdifferential* the set of all possible subgradients at $\mathbf{x} \in \mathbb{R}^n$.

Formally,

$$\partial f(\mathbf{x}) := \{\mathbf{s} \in \mathbb{R}^n : \mathbf{s} \text{ is a subgradient at } \mathbf{x}\}$$

Definition 22.0.3 (Descent direction). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ and let $\mathbf{x} \in \mathbb{R}^n$. We say that $\mathbf{d} \in \mathbb{R}^n$ is a *descent direction* if $\langle \mathbf{s}, \mathbf{d} \rangle < 0 \ \forall \mathbf{s} \in \partial f(\mathbf{x})$.

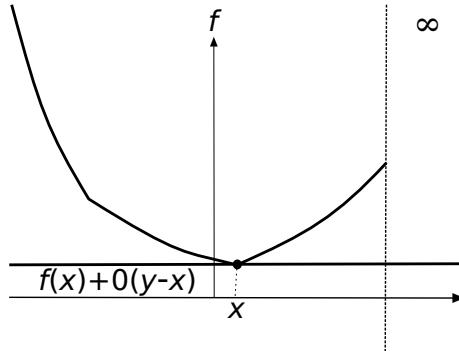


FIGURE 22.3: Pictorial example of subgradients of a non differentiable function.

Theorem 22.0.9. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$. \mathbf{x} global minimum $\iff 0 \in \partial f(\mathbf{x})$.

Notice that in general, when we are not in proximity of a border (where f is unbounded above) we get that the subdifferential is a compact interval.

Formally, $\partial f(\mathbf{x})$ closed and convex, compact $\forall \mathbf{x} \in \text{int dom}(f)$.

Moreover, we can prove the following

Fact 22.0.10. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$.

$$\partial f(\mathbf{x}) = \{\nabla f(\mathbf{x})\} \iff f \text{ differentiable at } \mathbf{x}$$

💡 Do you recall?

In Chapter 20, in Proposition 20.0.1 we derived the directional derivative from the gradient as $\frac{\partial f}{\partial \mathbf{d}}(\mathbf{x}) = \langle \nabla f(\mathbf{x}), \mathbf{d} \rangle$.

The following fact tries to provide a similar characterization of the subdifferential.

Fact 22.0.11. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ and let $\mathbf{x} \in \mathbb{R}^n$. If $\frac{\partial f}{\partial \mathbf{d}}(\mathbf{x}) = \sup\{\langle \mathbf{s}, \mathbf{d} \rangle : \mathbf{s} \in \partial f(\mathbf{x})\}$ then \mathbf{d} is a descent direction $\iff \langle \mathbf{s}, \mathbf{d} \rangle < 0 \ \forall \mathbf{s} \in \partial f(\mathbf{x})$.

As in the differentiable case, we are interested in moving in the steepest descent direction, formally $\mathbf{s}_* = -\text{argmin}\{\|\mathbf{s}\| : \mathbf{s} \in \partial f(\mathbf{x})\}$.

Example 22.0.1. Let us assume that we know the minimum point $\mathbf{x}_* \in \mathbb{R}^n$ for $f : \mathbb{R}^n \rightarrow \mathbb{R}$ and we are in \mathbf{x} at the current iteration. We would like to move towards \mathbf{x}_* using only the information provided by the subdifferential. $(-\partial f(\mathbf{x}))$ is convex and compact and all its elements $(-\mathbf{g}) \in \partial f(\mathbf{x})$ (subgradients) “point towards \mathbf{x}_* ”. Formally, $\langle \mathbf{g}, \mathbf{x}_* - \mathbf{x} \rangle < 0$.

Fact 22.0.12 (Linearity of subdifferential). Let $f, g : \mathbb{R}^n \rightarrow \mathbb{R}$ and take $\alpha, \beta \in \mathbb{R}_+$, then $\partial[\alpha f + \beta g](\mathbf{x}) = \alpha \partial f(\mathbf{x}) + \beta \partial g(\mathbf{x})$.

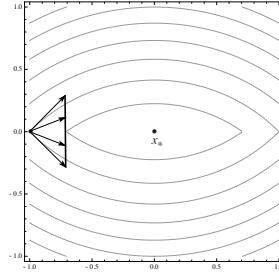


FIGURE 22.4: There are many different subgradients in \mathbf{x} . We pick the one with minimum norm among the ones which have a negative scalar product with $\mathbf{x} - \mathbf{x}_*$.

Fact 22.0.13 (Chain rule). *Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$.*

- Let $A \in M(n, \mathbb{R})$ and $\mathbf{b} \in \mathbb{R}^n$ then $\partial[f(A\mathbf{x} + \mathbf{b})] = A^T[\partial f](A\mathbf{x} + \mathbf{b})$;
- Let $g : \mathbb{R} \rightarrow \mathbb{R}$ be an increasing function. Then $\partial[g(f(\mathbf{x}))] = [\partial g](f(\mathbf{x}))[\partial f](\mathbf{x})$.

Definition 22.0.4 (ε -subgradient). *Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$. We say that $\mathbf{s} \in \mathbb{R}^n$ is ε -subgradient at $\mathbf{x} \in \mathbb{R}^n$ if it defines a support hyperplane passing ε below $\text{epi}(f)$. Formally,*

$$f(\mathbf{y}) \geq f(\mathbf{x}) + \mathbf{s}(\mathbf{y} - \mathbf{x}) - \varepsilon \quad \forall \mathbf{y} \in \mathbb{R}^n$$

Fact 22.0.14. *Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ and let $\mathbf{x} \in \mathbb{R}^n$. Saying that $\mathbf{0} \in \mathbb{R}^n$ is ε -subgradient for f in \mathbf{x} means that the function value $f(\mathbf{x})$ cannot be at distance grater than ε from the minimum $f(\mathbf{x}_*)$.*

Formally, $0 \in \partial_\varepsilon f(\mathbf{x}) \iff \mathbf{x}$ is ε -optimal.

We are now allowed to compute $\mathbf{s}_* = \operatorname{argmin}\{\|\mathbf{s}\| : \mathbf{s} \in \partial_\varepsilon f(\mathbf{x})\}$.

If $\mathbf{s}_* = \mathbf{0}$ then \mathbf{x} is ε optimal. Otherwise, $\exists \alpha > 0$ s.t. $f(\mathbf{x} - \alpha \mathbf{s}_*) \leq f(\mathbf{x}) - \varepsilon$ ($-\mathbf{s}_*$ is of ε -descent).

The ε -subgradient is very powerful, but the issue is that is even more expensive to compute than the subgradient.

Lecture 23

17th of October 2018

A. Frangioni

23.1 Optimization algorithms

💡 Do you recall?

We are interested in finding the minimum of a function through an iterative procedure, such that we start from an initial guess \mathbf{x}_0 and go on ($\mathbf{x}^i \rightsquigarrow \mathbf{x}^{i+1}$). We want to move towards the optimum.

Notice that with $\mathbf{x}^i \in \mathbb{R}^n$ we refer to the point \mathbf{x} at i -th iteration.

How to be sure we are in an optimum?

- (strong) $\{\mathbf{x}^i\} \rightarrow \mathbf{x}_*$: the whole sequence converges to an optimal solution;
- (weaker) all accumulation points of $\{\mathbf{x}^i\}$ are optimal solutions;
- (weakest) at least one accumulation point of $\{\mathbf{x}^i\}$ is optimal.

Such iterative process, can be held in two different ways:

LINE SEARCH: first choose $\mathbf{d}^i \in \mathbb{R}^n$ (direction), then choose $\alpha^i \in \mathbb{R}$ (that we term stepsize or equivalently “learning rate”) s.t. $\mathbf{x}^{i+1} \leftarrow \mathbf{x}^i + \alpha^i \mathbf{d}^i$;

TRUST REGION: first choose α^i (that we call *trust radius* and somehow indicates the ball in which we think that the information that we have is accurate), then choose \mathbf{d}^i .

In both these alternatives, it is crucial to choose a proper model to approximate function the objective function f .

23.1.1 Gradient method for quadratic functions

The simplest model we can build is a linear one, namely $L^i(\mathbf{x}) = L_{\mathbf{x}^i}(\mathbf{x}) = f(\mathbf{x}^i) + \nabla f(\mathbf{x}^i)(\mathbf{x} - \mathbf{x}^i)$ and find the descending direction of the model.

The first order information is accurate only around \mathbf{x}^i and this is why we should not move too far from \mathbf{x}^i , so we want $\alpha_i \rightarrow 0$ and then we pick the steepest descent direction, computed as $\mathbf{d}_i = \arg \min_d \left\{ \lim_{t \rightarrow 0} \frac{f(\mathbf{x}^i + t\mathbf{d})}{t} \right\} = -\nabla f(\mathbf{x}^i)$.

Notice that a too short step is bad either, because the gain in the value of the function is very little.

Alternatively, we can choose to pick a different step size at each iteration $\alpha^i = \arg \min_{\alpha} f(\mathbf{x}^i + \alpha \mathbf{d}^i)$ s.t. $\alpha \geq 0$.

The main drawback of choosing linear functions to approximate the objective function is that linear functions are unbounded below. The easiest family of functions which are still simple, yet more complex than linear ones are quadratic functions:

$$f(\mathbf{x}) = \frac{1}{2} \mathbf{x}^T Q \mathbf{x} + \mathbf{q}^T \mathbf{x}$$

where $Q \succeq 0$, otherwise f would be unbounded below.

The minimum is the point in which the gradient ($\nabla f(\mathbf{x}) = Q\mathbf{x} + \mathbf{q}$) is 0.

A formalization of the gradient descent algorithm for quadratic functions can be found in Algorithm 23.1.1.

ALGORITHM 23.1.1 Pseudocode for quadratic functions local minimum detection.

```

1: procedure SDQ( $f, \mathbf{x}, \varepsilon$ )
2:   while ( $\|\nabla f(\mathbf{x})\| > \varepsilon$ ) do
3:      $d \leftarrow -\nabla f(\mathbf{x});$ 
4:      $\alpha \leftarrow \frac{\|d\|^2}{(d^T Q d)};$ 
5:      $x \leftarrow \mathbf{x} + \alpha d;$ 
6:   end while
7: end procedure
```

As usual, we need to choose an error threshold ε under the one we feel allowed to think that the optimum has been reached and we set it as the norm of the gradient.

As far as the step size is concerned, let us find the minimum of $f(\mathbf{x}^i + \alpha \mathbf{d}^i)$ such

that $\alpha \geq 0$:

$$\begin{aligned}
\min_{\alpha} f(\mathbf{x}^i + \alpha \mathbf{d}^i) &= \min_{\alpha} \frac{1}{2} (\mathbf{x}^i + \alpha \mathbf{d}^i)^T Q (\mathbf{x}^i + \alpha \mathbf{d}^i) + \mathbf{q}^T (\mathbf{x}^i + \alpha \mathbf{d}^i) \\
&= \min_{\alpha} \frac{1}{2} \mathbf{x}^{i^T} Q \mathbf{x}^i + \frac{1}{2} \mathbf{x}^{i^T} Q \alpha \mathbf{d}^i + \frac{1}{2} \alpha \mathbf{d}^{i^T} Q \mathbf{x}^i + \frac{1}{2} \alpha \mathbf{d}^{i^T} Q \alpha \mathbf{d}^i + \mathbf{q}^T \mathbf{x}^i + \alpha \mathbf{q}^T \mathbf{d}^i \\
&\stackrel{(1)}{=} \min_{\alpha} \frac{1}{2} \mathbf{x}^{i^T} Q \mathbf{x}^i + \alpha \mathbf{x}^{i^T} Q \mathbf{d}^i + \frac{1}{2} \alpha^2 \mathbf{d}^{i^T} Q \mathbf{d}^i + \mathbf{q}^T \mathbf{x}^i + \alpha \mathbf{q}^T \mathbf{d}^i \\
&\stackrel{(2)}{=} \min_{\alpha} \alpha \mathbf{x}^{i^T} Q \mathbf{d}^i + \frac{1}{2} \alpha^2 \mathbf{d}^{i^T} Q \mathbf{d}^i + \alpha \mathbf{q}^T \mathbf{d}^i \\
&= \min_{\alpha} \left[\frac{1}{2} (\mathbf{d}^i)^T Q \mathbf{d}^i \right] \alpha^2 + \alpha [(\mathbf{x}^{i^T} Q + \mathbf{q}^T) \mathbf{d}^i] \\
&= \min_{\alpha} \left(\left(\frac{1}{2} \mathbf{d}^{i^T} Q \mathbf{d}^i \right) \alpha^2 + \alpha \cdot \left(\mathbf{x}^{i^T} Q + \mathbf{q}^T \right) \mathbf{d}^i \right)
\end{aligned}$$

where $\stackrel{(1)}{=}$ is due to the fact that Q is symmetric and $\stackrel{(2)}{=}$ follows from the fact that $\frac{1}{2} \mathbf{x}^{i^T} Q \mathbf{x}^i$ and $\mathbf{q}^T \mathbf{x}^i$ are constant with respect to α .

The minimum of such f can be found deriving and imposing the gradient to 0.

$$\begin{aligned}
\frac{\partial f(\mathbf{x}^i + \alpha \mathbf{d}^i)}{\partial \alpha} &= 0 \\
\frac{\partial \left(\left(\frac{1}{2} \mathbf{d}^{i^T} Q \mathbf{d}^i \right) \alpha^2 + \alpha \cdot \left(\mathbf{x}^{i^T} Q + \mathbf{q}^T \right) \mathbf{d}^i \right)}{\partial \alpha} &= 0 \\
\cancel{\alpha} \cdot \frac{1}{2} \mathbf{d}^{i^T} Q \mathbf{d}^i \cdot \alpha + (\mathbf{x}^{i^T} Q + \mathbf{q}^T) \mathbf{d}^i &= 0 \\
\mathbf{d}^{i^T} Q \mathbf{d}^i \cdot \alpha &= -(\mathbf{x}^{i^T} Q + \mathbf{q}^T) \mathbf{d}^i \\
\alpha &= -\frac{(\mathbf{x}^{i^T} Q + \mathbf{q}^T) \mathbf{d}^i}{\mathbf{d}^{i^T} Q \mathbf{d}^i}
\end{aligned}$$

Per me da qui non ha senso...

Hence, the minimum value of f is computed as

$$\begin{aligned}
f(\mathbf{x}^i + \alpha \mathbf{d}^i) &= \frac{1}{2} (\mathbf{x}^i + \alpha \mathbf{d}^i)^T Q (\mathbf{x}^i + \alpha \mathbf{d}^i) + \mathbf{q}^T (\mathbf{x}^i + \alpha \mathbf{d}^i) \\
&= \cancel{\frac{1}{2} \mathbf{x}^{i^T} Q \mathbf{x}^i} + (\mathbf{x}^i)^T Q (\alpha \mathbf{d}^i) + \frac{1}{2} \alpha^2 (\mathbf{d}^i)^T Q \mathbf{d}^i + \cancel{\mathbf{q}^T \mathbf{x}^i} + \alpha (\mathbf{q}^T \mathbf{d}^i) \\
&\stackrel{(1)}{=} \left(\frac{1}{2} (\mathbf{d}^i)^T Q \mathbf{d}^i \right) \alpha^2 + \alpha \left((\mathbf{x}^i Q + \mathbf{q}^T) \mathbf{d}^i \right) \\
&\stackrel{(2)}{=} \left(\frac{1}{2} (\mathbf{d}^i)^T Q \mathbf{d}^i \right) \alpha - \|\mathbf{d}^i\|
\end{aligned} \tag{23.1.1}$$

What if $(\mathbf{d}^i)^T Q \mathbf{d}^i = 0$? If Q is strictly positive definite this cannot happen, so the algorithm never breaks down.

... Fino a qui

At this point we are left with a few issues:

- Pick a proper $\varepsilon > 0$ (e.g. $\varepsilon = 0$ does not lead to convergence with machine precision)
- Check if $\{\mathbf{x}^i\}$ converges. This can be easily proved by construction:

$$\lim_{i \rightarrow \infty} \langle \nabla f(\mathbf{x}^i), \nabla f(\mathbf{x}^{i+1}) \rangle \stackrel{(*)}{=} \langle \nabla f(\mathbf{x}_*), \nabla f(\mathbf{x}_*) \rangle = 0$$

because at each step the gradients are orthogonal and $\stackrel{(*)}{=}$ follows from the fact that the gradient is continuous.

- Assess the convergence speed.

In general, showing how fast $\|\mathbf{x}^i - \mathbf{x}_*\|$ decreases is more involved than showing how fast $f(\mathbf{x}^i) - f_*$ decreases, but we do not know f_* . We concentrate on computing $\lim_{i \rightarrow \infty} \frac{f(\mathbf{x}^{i+1}) - f_*}{f(\mathbf{x}^i) - f_*} = R$. According to the values of p and R we get the following alternatives:

SUBLINEAR: $p = 1, R = 1$;

LINEAR: $p = 1, R < 1$;

SUPERLINEAR: $p = 1, R = 0$;

QUADRATIC: $p = 2, R > 0$.

Since the optimum is in $\mathbf{x}_* = -Q^{-1}\mathbf{q}$, we get that

$$\begin{aligned} f(\mathbf{x}_*) &= \frac{1}{2}(-Q^{-1}\mathbf{q})^T Q(-Q^{-1}\mathbf{q}) - \mathbf{q}^T Q^{-1}\mathbf{q} \\ &= \frac{1}{2}\mathbf{q}^T Q^{-1} Q Q^{-1}\mathbf{q} - \mathbf{q}^T Q^{-1}\mathbf{q} \\ &\stackrel{(*)}{=} \frac{1}{2}\mathbf{q}^T Q^{-1}\mathbf{q} - \mathbf{q}^T Q^{-1}\mathbf{q} \\ &= -\frac{1}{2}\mathbf{q}^T Q^{-1}\mathbf{q} \end{aligned}$$

Algebraically speaking, the minimum \mathbf{x}_* of a quadratic function is a stationary point i.e. $\nabla f(\mathbf{x}) = Q\mathbf{x} + \mathbf{q} = 0$ iff $\mathbf{x}_* = -Q^{-1}\mathbf{q}$ when Q is invertible. At this point, we need to assess the convergence speed of the algorithm, thus we need to define an error function \bar{f} , such that “the error at \mathbf{x} is the distance between \mathbf{x} and \mathbf{x}_* in $\|\cdot\|_Q$ ”:

$$\begin{aligned}
\bar{f}(\mathbf{x}) &= \frac{1}{2}(\mathbf{x} - \mathbf{x}_*)^T Q(\mathbf{x} - \mathbf{x}_*) \\
&= \frac{1}{2}\mathbf{x}^T Q\mathbf{x} + \frac{1}{2}\mathbf{x}_*^T Q\mathbf{x}_* - \mathbf{x}^T(Q\mathbf{x}_*) \\
&\stackrel{(2)}{=} \frac{1}{2}\mathbf{x}^T Q\mathbf{x} + \frac{1}{2}(Q^{-1}\mathbf{q})^T Q(Q^{-1}\mathbf{q}) - \mathbf{x}^T \cancel{Q}(-Q^{-1}\mathbf{q}) \\
&= \frac{1}{2}\mathbf{x}^T Q\mathbf{x} + \frac{1}{2}Q^{-1}\mathbf{q}^T Q(Q^{-1}\mathbf{q}) + \mathbf{q}^T \mathbf{x} \\
&= \frac{1}{2}\mathbf{x}^T Q\mathbf{x} + \frac{1}{2}\mathbf{q}^T \cancel{Q} \cancel{Q}^{-1}\mathbf{q} + \mathbf{q}^T \mathbf{x} \\
&= \frac{1}{2}\mathbf{x}^T Q\mathbf{x} + \frac{1}{2}\mathbf{q}^T Q^{-1}\mathbf{q} + \mathbf{q}^T \mathbf{x} \\
&= \frac{1}{2}\mathbf{x}^T Q\mathbf{x} + \mathbf{q}^T \mathbf{x} + \frac{1}{2}\mathbf{q}^T Q^{-1}\mathbf{q} \\
&= f(\mathbf{x}) - f(\mathbf{x}_*)
\end{aligned} \tag{23.1.2}$$

where $\stackrel{(2)}{=}$ follows from the equality $\mathbf{x}_* = -Q^{-1}\mathbf{q}$.

Rispetto alle slide, qui denoto con $\bar{f}(\cdot)$ la funzione che quantifica l'errore, anzichè chiamarla $f_*(\cdot)$, perché questa seconda notazione si confonde con f_* (senza argomento), che è il valore ottimo della funzione obiettivo.

Lecture 24

19th of October 2018

A. Frangioni

24.0.1 Gradient method for quadratic functions (cont.)

💡 Do you recall?

The family of quadratic functions is the simplest possible family of functions where a minimum exists. A quadratic function is defined as: $f(\mathbf{x}) = \frac{1}{2}\mathbf{x}^T Q\mathbf{x} + \mathbf{q}^T \mathbf{x}$, so its gradient is computed as $\nabla f(\mathbf{x}) = Q\mathbf{x} + \mathbf{q}$. We are interested in finding local minima of the function f .

In last lecture we were left with analyzing the convergence of the algorithm

Fact 24.0.1. *A quadratic function $f : \mathbb{R}^n \rightarrow \mathbb{R}^n$, s.t. $f(\mathbf{x}) = \frac{1}{2}\mathbf{x}^T Q\mathbf{x} + \mathbf{q}^T \mathbf{x}$ admits a minimum iff $Q \succeq 0$ (Q is positive semidefinite).*

The gradient method generates points that move along orthogonal directions. More formally, $\mathbf{x}^{i+1} = \mathbf{x}^i + \alpha^i \mathbf{d}^i$, where $\mathbf{d}^i = -\nabla f(\mathbf{x}^i) = -Q\mathbf{x}^i - \mathbf{q}$

Fact 24.0.2. *Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ be a quadratic function and let us consider the gradient descent algorithm for minimizing f , where at iteration i the moving direction $\mathbf{d}^i \in \mathbb{R}^n$ and the step size $\alpha^i \in \mathbb{R}$ are computed. The following holds: $\langle \mathbf{d}^i, \mathbf{d}^{i+1} \rangle = 0$.*

Proof. Let us rewrite \mathbf{d}^{i+1} in terms of \mathbf{d}^i :

$$\begin{aligned}\mathbf{d}^{i+1} &= -Q\mathbf{x}^{i+1} - \mathbf{q} \\ &= -Q(\mathbf{x}^i + \alpha^i \mathbf{d}^i) - \mathbf{q} \\ &= -Q\mathbf{x}^i - Q\alpha^i \mathbf{d}^i - \mathbf{q} \\ &= \mathbf{d}^i - Q\alpha^i \mathbf{d}^i\end{aligned}$$

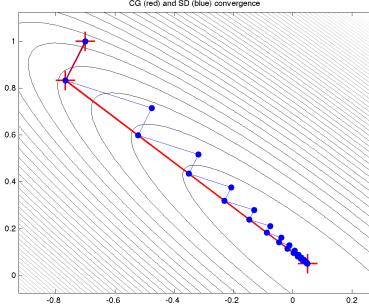


FIGURE 24.1: Some iterations of the gradient method

If we plug this into the scalar product $\langle \mathbf{d}^i, \mathbf{d}^{i+1} \rangle$ we get

$$\begin{aligned} \langle \mathbf{d}^i, \mathbf{d}^{i+1} \rangle &= \langle \mathbf{d}^i, \mathbf{d}^i - Q\alpha^i \mathbf{d}^i \rangle \\ &= \|\mathbf{d}^i\|^2 - \mathbf{d}^i{}^T Q\alpha^i \mathbf{d}^i \\ &= \|\mathbf{d}^i\|^2 - \alpha^i \mathbf{d}^i{}^T Q \mathbf{d}^i \\ &\stackrel{*}{=} \|\mathbf{d}^i\|^2 - \frac{\|\mathbf{d}^i\|^2}{\mathbf{d}^i{}^T Q \mathbf{d}^i} \cdot \cancel{\mathbf{d}^i{}^T Q \mathbf{d}^i} \end{aligned}$$

where $\stackrel{*}{=}$ follows from $\alpha^i := \frac{\|\mathbf{d}^i\|^2}{\mathbf{d}^i{}^T Q \mathbf{d}^i}$. \square

Let us go back to the error function

$$\bar{f}(\mathbf{x}^{i+1}) = f(\mathbf{x}^{i+1}) - f_* = \frac{1}{2} (\mathbf{x}^{i+1} - \mathbf{x}_*)^T Q (\mathbf{x}^{i+1} - \mathbf{x}_*)$$

The $(i + 1)$ th iteration is computed as

$$\begin{aligned} \mathbf{x}^{i+1} &= \mathbf{x}^i + \alpha^i \mathbf{d}^i \\ &\stackrel{(a)}{=} \mathbf{x}^i + \frac{\|\mathbf{d}^i\|^2}{\mathbf{d}^i{}^T Q \mathbf{d}^i} \mathbf{d}^i \\ &\stackrel{(b)}{=} \mathbf{x}^{i+1} = \mathbf{x}^i + \frac{\|-\mathbf{Q}\mathbf{x}^i - \mathbf{q}\|^2}{(-\mathbf{Q}\mathbf{x}^i - \mathbf{q})^T Q (-\mathbf{Q}\mathbf{x}^i - \mathbf{q})} \end{aligned}$$

where $\stackrel{(a)}{=}$ follows from the definition $\alpha^i := \frac{\|\mathbf{d}^i\|^2}{\mathbf{d}^i{}^T Q \mathbf{d}^i}$ and $\stackrel{(b)}{=}$ from the fact that the moving direction is the opposite of the gradient (formally, $\mathbf{d}^i = -\mathbf{Q}\mathbf{x}^i - \mathbf{q}$).

This means not only **linear convergence**, but a bit more, because linear convergence only takes into consideration steps in proximity to the limit, while this formula holds at the beginning as well.

Fact 24.0.3. Let $A \in M(n, \mathbb{R})$ be a positive semidefinite matrix. $\forall \mathbf{v} \in \mathbb{R}^n$ $\mathbf{v}^T A \mathbf{v} = \|\mathbf{v}\|_A^2$.

Fact 24.0.4. If Q is positive definite we can say that the error goes like $1 - \left(\frac{\|\mathbf{d}_i\|^4}{(\mathbf{d}_i^T Q \mathbf{d}_i)(\mathbf{d}_i^T Q^{-1} \mathbf{d}_i)} \right)$ or, equivalently, $1 - \frac{\|\mathbf{d}_i\|_I^2}{\|\mathbf{d}_i\|_Q^2} \cdot \frac{\|\mathbf{d}_i\|_I^2}{\|\mathbf{d}_i\|_{Q^{-1}}^2}$.

Fare proof, $\bar{f}(\mathbf{x}^{i+1}) = 1 - \left(\frac{\|\mathbf{d}_i\|^4}{(\mathbf{d}_i^T Q \mathbf{d}_i)(\mathbf{d}_i^T Q^{-1} \mathbf{d}_i)} \right) \bar{f}(\mathbf{x}^i)$ (hint: for $y^i = x^i - x_*$, $d^i = Qy^i$)

We are measuring a vector with three different norms. We would like to estimate $\frac{\langle \mathbf{d}_i, \mathbf{d}_i \rangle}{\mathbf{d}_i^T Q \mathbf{d}_i}$.

Fact 24.0.5. Let $A \in M(n, \mathbb{R})$ be a positive semidefinite matrix. Given λ_i the eigenvalues of A , $\frac{1}{\lambda_i}$ are the eigenvalues of the matrix A^{-1} .

We can say that $\lambda_n \|\mathbf{x}\|^2 \leq \mathbf{x}^T Q \mathbf{x} \leq \lambda_1 \|\mathbf{x}\|^2$, where λ_n is the smallest eigenvalue, while λ_1 is the biggest.

We are looking for a close formula for calculating the convergence rate, since it depends recursively by the steps done. So we want to do a worst case analysis in order to find a faster way to calculate convergence rate.

We want to prove that R is smaller than 1 so we are looking for an upper-bound.

A coarse upperbound is $(1 - \frac{\lambda_n}{\lambda_1})$, but we can prove more:

Fact 24.0.6. Let $A \in M(n, \mathbb{R})$ be a positive definite matrix. Given λ_i the eigenvalues of A ,

$$\forall \mathbf{v} \in \mathbb{R}^n \quad \frac{\|\mathbf{v}\|^4}{(\mathbf{v}^T A \mathbf{v})(\mathbf{v}^T A^{-1} \mathbf{v})} \geq \frac{4\lambda^1 \lambda^n}{(\lambda^1 + \lambda^n)^2}$$

We won't see the proof of this fact.

R close to 0 means that the algorithm is converging fast, so when the larger eigenvalue (λ_1) and the smallest eigenvalue (λ_n) are very close to each-other the algorithm is converging fast. We can say that, since the eigenvalues represent the length of the axis of the ellipsoids (level sets), the algorithm is converging fast when the ellipsoid has a round shape.

We can provide an even better estimate, which is $(\frac{\lambda_1 - \lambda_n}{\lambda_1 + \lambda_n})^2$. At this point, provided that we can estimate the number of iterations needed to reach convergence, how can we say that such a number is a good result? As usual, it depends on how that number is obtained. If it is dimensional independent it is very good, because it scales well (when the size of the space - number of variables - increases). It only depends on the conditioning of the matrix Q .

Of course as n grows Q changes, so in practice it may happen that the conditioning of the problem is worsening as n grows.

If the balls are very rounded the zig zags needed to start converging are very few.

Note

So far, we provided a bound for the convergence speed of the algorithm when Q is positive definite. What can we say when Q is positive semidefinite? The algorithm works, but we can't provide an upper-bound for the convergence rate. We are even more restrictive when dealing with machine precision, since if there is an eigenvalue which is bigger than zero, but very close to zero, it turns out to be 0 on the machine, so we cannot provide an upper-bound. We will see how to deal with this case.

TODO: mettere riferimento

24.0.2 Quadratic gradient descent in Matlab

When we are trying to move from theory to practice we need to take into account some practical issue, such as picking a good value for ε . A good idea would be the norm of the gradient, but in order not to end up in a loop we need to provide also some performances bound, e.g. the maximum number of iterations or the maximum amount of time.

Let us suppose that we implemented the quadratic gradient descent algorithm in Matlab, with the following signature:

```
1 function [x, status] = SDQ(Q, q, x, fStar, eps, MaxIter)
```

where `fStar` is the optimal value, which is used to provide an estimate of the convergence.

Note

When implementing functions for optimization is important to:

- check the coherence of input values (namely, if the user passed allowed parameters);
- give all possible information about your result (for example if the algorithm stopped because the maximum number of iterations was reached or because the epsilon value was reached);
- check at any iteration the values of variables during computations (e.g. before dividing by a quantity that may be smaller than the precision check it before computing the ratio);
- design a good log, in order to understand what is going on at each step.

Example 24.0.1. In the implementation of `SDQ`* provided by professor Frangioni, the execution keeps track of the actual ratio between the error at one step and

*<https://elearning.di.unipi.it/mod/resource/view.php?id=5033>

the error at the next step. This information provides insights about how many orders of magnitude the error decreases. We can find starting points where the ratio is exactly R . We can observe that when the conditioning is quite good the error decrease faster than the R limitation, but as soon as we change the values for Q and q things may change completely.

If the reader wants to run some examples, he should notice that if the conditioning grows, the number of iterations needed to find the minimum increases as well.

Trying the algorithm on some examples shows that the theoretical results are reflected well in the practical case.

Error

As mentioned above, if we picked as starting point for the algorithm a value that does not lead to a minimizing sequence, we need to stop iterating after a while, possibly when we are close enough to the solution, i.e. the absolute error ε_A is small enough: $\varepsilon_A = |f(x^i) - f_*| \leq \varepsilon$. It is in this context that we introduce the concept of **relative error**, that compares the error with the value of the function.

Definition 24.0.1 (Relative error). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ be a quadratic function and let us consider the gradient descent algorithm for minimizing f , where at iteration i the moving direction $\mathbf{d}^i \in \mathbb{R}^n$ and the step size $\alpha^i \in \mathbb{R}$ are computed. We term **relative error** at the i -th iteration with respect to the optimal value and denote ε_R

$$\varepsilon_R = |f(x^i) - f_*| / |f_*| = \varepsilon_A / |f_*| \leq \varepsilon$$

This relative error is invariant for scaling transformations. Notice that if f^* might be zero the formula should be changed.

In practice, computing such error is unfeasible because we do not know f^* . In this very common case we can substitute f^* with a good lower bound $\underline{f} \leq f^*$ for f^* , but in this course we will not focus on finding \underline{f} .

Alternative stopping conditions are bounds on the norm of the gradient:

- $\|\nabla f(x^i)\| \leq \varepsilon$ (“absolute version”)
- $\|\nabla f(x^i)\| / \|\nabla f(x^0)\| \leq \varepsilon$ (“relative version”)

The second stopping condition is expressed in relation to the value of the norm of the gradient at the starting point.

We usually choose the norm of the gradient as a threshold for precision, but we do not know how this quantity relates to ε_A or ε_B .

Example 24.0.2. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ a convex function and let $X = \mathcal{B}(\mathbf{0}, r)$. Estimate ε_A when $\|\nabla f(\mathbf{x}^i)\| \leq \varepsilon$. Provided that we are studying a convex function we can easily find the minimum in a ball. We can then minimize the linear function in the range of the ball and that minimum is surely a lower bound.

Lecture 25

24th of October 2018

A. Frangioni

As we said so far, dealing with quadratic (approximations of) functions is pretty straightforward. In this chapter and a few subsequent ones, we will explore functions that have different shapes.

25.1 Gradient method for non quadratic functions

💡 Do you recall?

Let us consider that we are sitting in a point \mathbf{x}^i at the i -th iteration. In order to compute the next step, in the case of quadratic functions, we decided to put $\alpha^i = \frac{\|\mathbf{d}^i\|^2}{\mathbf{d}^{iT} Q \mathbf{d}^i}$ and $\mathbf{d}^i = -\nabla f(\mathbf{x}^i) = -Q\mathbf{x}^i - \mathbf{q}$. We may recall from last lecture that the proof of the orthogonality of the gradient (Proposition 24.0.2) does not depend on the quadratic nature of our functions, so it could work in the non-quadratic case as well.

We would like to find a more general form for the step size, which does not depend on the fact that the function is quadratic.

The algorithm for finding local minima of non quadratic functions has the same structure of the one used for quadratic ones, i.e. first compute the direction of the step and then compute its size and it is formalized in Algorithm 25.1.1.

ALGORITHM 25.1.1 Pseudocode for non quadratic functions local minimum detection.

```

1: procedure SDQ( $f, \mathbf{x}, \varepsilon$ )
2:   while ( $\|\nabla f(\mathbf{x})\| > \varepsilon$ ) do
3:      $\mathbf{d} \leftarrow -\nabla f(\mathbf{x});$ 
4:      $\alpha \leftarrow \frac{\|\mathbf{d}\|^2}{(\mathbf{d}^T Q \mathbf{d})};$ 
5:      $\mathbf{x} \leftarrow \mathbf{x} + \alpha \mathbf{d};$ 
6:   end while
7: end procedure
```

We will see that, differently from the quadratic case (where the gradient was $\nabla f(\mathbf{x}) = Q\mathbf{x} + \mathbf{q}$) computing the gradient in this more general case is not trivial. We are now interested in assessing the converge speed of Algorithm 25.1.1.

Theorem 25.1.1 (Convergence speed). *Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ be a function in \mathcal{C}^2 and let \mathbf{x}_* be a local minimum for f such that the Hessian of f is strictly positive definite ($\nabla^2 f(\mathbf{x}_*) \succ 0$). If $\{\mathbf{x}^i\}$ converges it is a minimizing sequence. Formally,*

$$\{\mathbf{x}^i\} \rightarrow \mathbf{x}_* \implies \{f(\mathbf{x}^i)\} \rightarrow f(\mathbf{x}_*)$$

linearly, with same R as the quadratic case, depending on λ_1 and λ_n , respectively the biggest and smallest eigenvalues of the Hessian.

This theorem means that if the function is differentiable and the Hessian is strictly positive definite then when getting closer and closer to the minimum, the function is more and more similar to a quadratic function.

This similarity is a good new, since we can use the same methods of the quadratic case, but, as usual, we must pay attention to **conditioning**.

25.1.1 Finding the best step size

At this point we are left with the task of computing a step length, say α^i such that $f(\mathbf{x}^i + \alpha^i \mathbf{d}^i) < f(\mathbf{x}^k)$. This may necessitate a one-dimensional (or line) search, i.e. finding the local minimum of the one dimensional function φ^i , s.t.

$$\varphi^i(\alpha) = f(\mathbf{x}^i + \alpha \mathbf{d}^i),$$

where $\mathbf{d}^i = -\nabla f(\mathbf{x}^i)$.

Let us omit the index i , since we are concentrating on a single iteration. We are interested in finding $\alpha^* \in \mathbb{R}$ such that $\varphi'(\alpha^*) = 0$

Example 25.1.1. *Let us take $f : \mathbb{R}^2 \rightarrow \mathbb{R}$ s.t. $f(x_1, x_2) = x_1^2 e^{x_2}$. We can differentiate f and obtain $\nabla f(x_1, x_2) = (2x_1 e^{x_2}, x_1^2 e^{x_2})^T$.*

Let us suppose that at the i -th iteration $\mathbf{x} = (0, 1)^T$, so $\nabla f(1, 0) = (2, 1)$. Now $\mathbf{x}(\alpha) = \mathbf{x} - \alpha \nabla f(\mathbf{x}) = (1, 0)^T - \alpha(2, 1)^T = (1 - 2\alpha, 0 - \alpha)^T$.

At this point we obtain $\varphi(\alpha) = f(\mathbf{x}(\alpha)) = (1 - 2\alpha)^2 e^{-\alpha}$.

In general, we are not interested in finding those points where $\varphi'(\alpha) = 0$, instead we want to force the directional derivative to be small (by picking a threshold ε' s.t. $|\varphi'(\alpha)| \leq \varepsilon'$).

Fact 25.1.2. Let $\varphi : \mathbb{R} \rightarrow \mathbb{R}$, such that $\varphi(\alpha) = f(\mathbf{x}^i + \alpha \mathbf{d}^i)$, $\varphi'(\alpha) = \langle \nabla f(\mathbf{x}^i + \alpha \mathbf{d}^i), \mathbf{d}^i \rangle$.

Proof.

$$\frac{\partial f(\mathbf{x}^i + \alpha \mathbf{d}^i)}{\partial \alpha} = \frac{\partial f}{\partial \mathbf{x}^i + \alpha \mathbf{d}^i} \cdot \frac{\partial \mathbf{x}^i + \alpha \mathbf{d}^i}{\partial \alpha} = \nabla f(\mathbf{x}^i + \alpha \mathbf{d}^i) \cdot \mathbf{d}^i$$

□

Fact 25.1.3. We claim that $\varepsilon' = \varepsilon$.

Proof. We want to find the relationship between the two parameters ε and ε' . Assuming that we've got a black box that finds α , given ε' , we are interested in computing ε' .

Key idea: Normalization of the direction.

We may normalize the direction of movement \mathbf{d}^i without perturbing the behaviour of the algorithm: $\mathbf{d}^i = -\frac{\nabla f(\mathbf{x}^i)}{\|\nabla f(\mathbf{x}^i)\|}$. Note that we're not worried of dividing by the norm of the gradient, since if it gets 0 we have already stopped the procedure.

In this new context $\|\mathbf{d}^i\| = 1$ and

$$\varphi'(0) = \frac{\partial f(\mathbf{x})}{\partial \mathbf{d}}$$

(From Proposition 25.1.2) $= \langle \nabla f(\mathbf{x}), \mathbf{x} \rangle$

$$\begin{aligned} &= \langle \nabla f(\mathbf{x}), \frac{-\nabla f(\mathbf{x})}{\|\nabla f(\mathbf{x})\|} \rangle \\ &= -\frac{\langle \nabla f(\mathbf{x}), \nabla f(\mathbf{x}) \rangle}{\|\nabla f(\mathbf{x})\|} \\ &= -\frac{\|\nabla f(\mathbf{x})\|^2}{\|\nabla f(\mathbf{x})\|} \\ &= -\|\nabla f(\mathbf{x})\| \end{aligned}$$

$$|\varphi'(\alpha^i)| = |\langle \nabla f(\mathbf{x}^i + \alpha \mathbf{d}^i), \mathbf{d}^i \rangle| = |\langle \nabla f(\mathbf{x}^{i+1}), \mathbf{d}^i \rangle| = \left| \langle \nabla f(\mathbf{x}^{i+1}), -\frac{\nabla f(\mathbf{x}^i)}{\|\nabla f(\mathbf{x}^i)\|} \rangle \right|$$

If the iterations converge to the optimum point (formally, $\{\mathbf{x}^i\} \rightarrow \mathbf{x}$)

$$\begin{aligned} \lim_{i \rightarrow \infty} |\varphi'(\alpha^i)| &= \lim_{i \rightarrow \infty} \left| \langle \nabla f(\mathbf{x}^{i+1}), \frac{\nabla f(\mathbf{x}^i)}{\|\nabla f(\mathbf{x}^i)\|} \rangle \right| \\ &= \left| \langle \nabla f(\mathbf{x}), \frac{\nabla f(\mathbf{x})}{\|\nabla f(\mathbf{x})\|} \rangle \right| \\ &= \left| \frac{\langle \nabla f(\mathbf{x}), \nabla f(\mathbf{x}) \rangle}{\|\nabla f(\mathbf{x})\|} \right| \\ &= \|\nabla f(\mathbf{x})\| \leq \varepsilon' \end{aligned}$$

At each iteration of the gradient descent algorithm we check the condition $\|\nabla f(\mathbf{x}^i)\| > \varepsilon$, hence we have the thesis. \square

Per each phase the new epsilon is obtained $\varepsilon \leftarrow \varepsilon \|\nabla f(\mathbf{x}^i)\|$.

If we can prove that the algorithm is converging we know when to stop.

This convergence is not the perfect mathematical convergence, since $\varepsilon \neq 0$, because the line search will never terminate.

In the rest of this subsection we will propose three different kinds of algorithms for performing the line search on φ :

- First order algorithms;
- Second order algorithms;
- Zero order algorithms;
- Inexact line search;

where the first three approaches provide an exact solution, while the fourth one is less accurate but faster.

First order algorithms

This class of algorithms makes use of the first derivative of the function φ for finding the exact minimum of φ , namely we want to find the minimum points of φ , which corresponds to points where the first order derivative is zero and it goes from negative to positive.

We will present three different algorithms for this class, two of them are iterative, while the last one is a direct method. For the first two methods, we would like to reduce the range in which performing the search, at each step.

Doubling: How can we be sure that in a given range there is a point where the derivative is 0? Rolle's theorem, as shown in Figure 25.1.

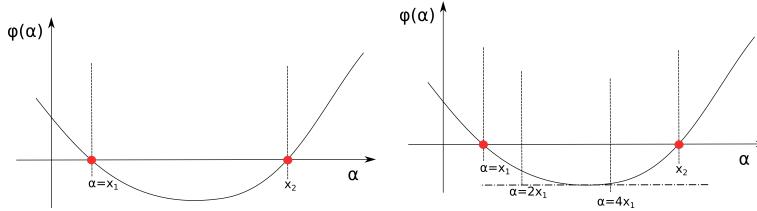


FIGURE 25.1: First, we restrict from \mathbb{R} to $[x_1, x_2]$, then double α until the derivative is greater than 0

Since the gradient is continuous the directional derivative is continuous, so φ is continuous (the scalar product is continuous).

Actually, we only need to find where the derivative is positive, because the 0 of the derivative is between the previous value and this point. A formalization of the algorithm can be found in Algorithm 25.1.2.

ALGORITHM 25.1.2 First algorithm for exact line search

```
1:  $\bar{\alpha} \leftarrow x_1$ ; #or whatever value > 0
2: while ( $\varphi'(\bar{\alpha}) < 0$ ) do
3:    $\bar{\alpha} \leftarrow 2\bar{\alpha}$ ; #or whatever factor > 1
4: end while
```

Since we are dealing with machine precision we stop when $\alpha < -10^{308}$, which is the smallest value for a double.

Works if φ **coercive**: $\lim_{\alpha \rightarrow \infty} \varphi(\alpha) = \infty$ (ex. f strongly convex)

Exercise 25.1.1. Build an example where $\bar{\alpha}$ exists but it is not found by this algorithm.

Solution: The function changes its derivative in a range between α and 2α .

Bisection: we pick the middle point of the interval $[\alpha_-, \alpha_+]$, as shown in Algorithm 25.1.3.

ALGORITHM 25.1.3 Bisection algorithm

```
1: procedure LSBM( $(\varphi', \bar{\alpha}, \varepsilon)$ )
2:    $\alpha_- \leftarrow 0$ ;
3:    $\alpha_+ \leftarrow \bar{\alpha}$ ;
4:    $\alpha \leftarrow \alpha_+$ ;
5:   while ( $|\varphi'(\alpha)| > \varepsilon$ ) do
6:      $\alpha \leftarrow (\alpha_+ + \alpha_-)/2$ ;
7:     if ( $\varphi'(\alpha) < 0$ ) then
8:        $\alpha_- \leftarrow \alpha$ ;
9:     else
10:     $\alpha_+ \leftarrow \alpha$ ;
11:   end if
12:   end while
13: end procedure
```

Quadratic approximation: we use the first order information to build a quadratic model m that approximates φ in a ball centered in α .

We may use the information we have about the function, since we know $\varphi(\alpha^-)$, $\varphi'(\alpha^-)$, $\varphi(\alpha^+)$ and $\varphi'(\alpha^+)$.

At this point we can write a model ($m(\alpha) = a\alpha^2 + b\alpha + c$) and specialize it with the information we have, via computing a linear system:

$$\begin{cases} a(\alpha^-)^2 + b(\alpha^-) + c = \varphi(\alpha^-) \\ a(\alpha^+)^2 + b(\alpha^+) + c = \varphi(\alpha^+) \\ 2a\alpha^- + b = \varphi'(\alpha^-) \\ 2a\alpha^+ + b = \varphi'(\alpha^+) \end{cases}$$

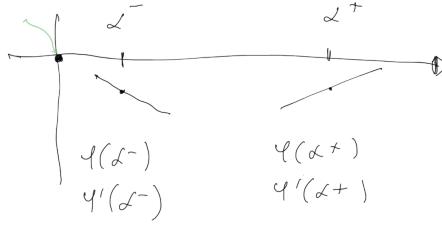


FIGURE 25.2: The information we have about function φ

Then we look for the stationary point of this function and that's the ball where I'm likely to find the root of the derivative.

Fact 25.1.4. *Let $\varphi : \mathbb{R} \rightarrow \mathbb{R}$ such that $\varphi \in C^3$, then quadratic interpolation has convergence of order $1 < p < 2$ (superlinear).*

In Figure 25.3 we can observe a situation in which the hypotheses of Proposition 25.1.4 are not satisfied.

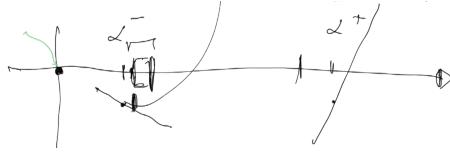


FIGURE 25.3: If the function is not C^3 and one derivative is very big, then the range does not shrink much.

We would like to modify the formula to have at least linear convergence.

We can ensure to move not too close to one of the extremes, for example more than 10%.

It is possible to build a more complex model, i.e. a cubic function. Although this is pretty involved, it allows the convergence to get quadratic, which is better than linear.

Lecture 26

25th of October 2018

A. Frangioni

💡 Do you recall?

During last lecture we started dealing with algorithms for performing a line search for the minimum of φ , in particular we introduced first order algorithms.

Second order algorithms

Theorem 26.0.1. Given $f : \mathbb{R}^n \rightarrow \mathbb{R}$ such that $f \in \mathcal{C}^2$ and let $\varphi(\alpha) = \mathbf{x} + \alpha\mathbf{d}$, $\exists \varphi''(\alpha) = \mathbf{d}^T \nabla^2 f(\mathbf{x} + \alpha\mathbf{d}) \mathbf{d}$ and it is continuous.

Proof.

$$\begin{aligned}\frac{\partial \langle \nabla f(\mathbf{x} + \alpha\mathbf{d}), \mathbf{d} \rangle}{\partial \alpha} &= \frac{\partial \langle \mathbf{d}, \nabla f(\mathbf{x} + \alpha\mathbf{d}) \rangle}{\partial \alpha} \\ &= \frac{\partial (\nabla f(\mathbf{x} + \alpha\mathbf{d}) \cdot \mathbf{d})}{\partial \alpha} \\ &= \mathbf{d} \cdot \left(\frac{\partial \nabla f(\mathbf{x} + \alpha\mathbf{d})}{\partial \alpha} \cdot \frac{\partial (\mathbf{x} + \alpha\mathbf{d})}{\partial \alpha} \right) \\ &= \mathbf{d}^T \nabla^2 f(\mathbf{x} + \alpha\mathbf{d}) \cdot \mathbf{d}\end{aligned}$$

□

Since we are looking for a point where the derivative $\varphi'(\alpha) = 0$, we may use the second derivative to write a model and, assuming to trust the model, it can be studied.

Definition 26.0.1 (Model–Newton tangent method). *Our model, in this case is*

$$\varphi'(\alpha) \approx \varphi'(\alpha^k) + \varphi''(\alpha^k)(\alpha - \alpha^k) = 0 \text{ iff } \alpha = \alpha^k - \varphi'(\alpha^k)/\varphi''(\alpha^k)$$

In this context, solving $\varphi'(\alpha) = 0$ implies finding those α such that $\alpha = \alpha^k - \varphi'(\alpha^k)\varphi''(\alpha^k)$

Spiegazione: sono nel punto α^k e stimo $\varphi'(\alpha)$ come $\varphi'(\alpha^k) + \varphi''(\alpha^k)[\text{coefficiente angolare}] \cdot (\alpha - \alpha^k)$

ALGORITHM 26.0.1 Pseudocode for Newton method.

```

1: procedure LSNM( $\varphi'$ ,  $\varphi''$ ,  $\alpha$ ,  $\varepsilon$ )
2:   while ( $|\varphi'(\alpha)| > \varepsilon$ ) do
3:      $\alpha \leftarrow \alpha - \frac{\varphi'(\alpha)}{\varphi''(\alpha)}$ ;
4:   end while
5: end procedure

```

We need to understand when and why $\varphi''(\alpha) \neq 0$ and when and why this method converges. The following theorem formalizes the fact that if we start from a point α^0 which is close to the optimum we reach the optimum with quadratic speed.

Theorem 26.0.2. *Let $\varphi : \mathbb{R}^n \rightarrow \mathbb{R}$ and let $\varphi \in C^3$ and take $\alpha_* \in \mathbb{R}$ such that $\varphi'(\alpha_*) = 0$ and $\varphi''(\alpha_*) \neq 0$. $\exists \delta > 0$ s.t. if $\alpha^0 \in [\alpha_* - \delta, \alpha_* + \delta]$ then $\{\alpha^k\} \rightarrow \alpha_*$, with $p = 2$.*

Proof. We are in the hypothesis that the function φ is three times differentiable and we would like to prove that $\alpha^{k+1} - \alpha_* \rightarrow 0$

We want to compute how much the error is, if compared to the error at the previous iteration. Since (1) $\alpha^{k+1} := \alpha^k - \frac{\varphi'(\alpha^k)}{\varphi''(\alpha^k)}$ and (2) $\varphi'(\alpha_*) = 0$ and we get

$$\begin{aligned}
\alpha^{k+1} - \alpha_* &\stackrel{(1)}{=} \alpha^k - \frac{(\varphi'(\alpha^k))}{\varphi''(\alpha^k)} - \alpha_* \\
&= \alpha^k - \alpha_* - \frac{(\varphi'(\alpha^k))}{\varphi''(\alpha^k)} \\
&\stackrel{(2)}{=} \alpha^k - \alpha_* - \frac{(\varphi'(\alpha^k) - \varphi'(\alpha_*))}{\varphi''(\alpha^k)} \\
&= \frac{-\varphi''(\alpha^k) \cdot (\alpha_* - \alpha_k) - \varphi'(\alpha^k) + \varphi'(\alpha_*)}{\varphi''(\alpha^k)}
\end{aligned}$$

Let us write the second form of the first order Taylor's model centered in α^k (Definition 20.0.6, Equation (20.0.1))

$$\exists \beta \in [\alpha^k, \alpha_*] \text{ s.t. } \varphi'(\alpha_*) = \varphi'(\alpha^k) + \varphi''(\alpha^k)(\alpha_* - \alpha^k) + \varphi'''(\beta) \frac{(\alpha_* - \alpha^k)^2}{2}$$

and plug it into the above equation:

$$\begin{aligned}
\alpha^{k+1} - \alpha_* &= \frac{-\varphi''(\alpha^k) \cdot (\alpha_* - \alpha_k) - \varphi'(\alpha^k) + \varphi'(\alpha_*)}{\varphi''(\alpha^k)} \\
&= \frac{-\cancel{\varphi''(\alpha^k)} \cdot \cancel{(\alpha_* - \alpha_k)} - \cancel{\varphi'(\alpha^k)} + \cancel{\varphi'(\alpha^k)} + \cancel{\varphi''(\alpha^k)} \cdot \cancel{(\alpha_* - \alpha_k)} + \frac{1}{2}\varphi'''(\beta) \cdot (\alpha_* - \alpha^k)}{\varphi''(\alpha^k)} \\
&= \frac{\frac{1}{2}\varphi'''(\beta) \cdot (\alpha_* - \alpha^k)}{\varphi''(\alpha^k)}
\end{aligned}$$

We can say that the quantity $2\varphi''(\alpha^k)$ does not become too small and that the numerator $\varphi'''(\beta)$ does not become too big. This is proved since $\exists \delta > 0$ s.t. $\varphi''(\alpha) \geq k_2 > 0$ and also $|\varphi'''(\beta)| \leq k_1 < \infty$. We can go on bounding the difference between α^{k+1} and α_* as follows: for $\alpha, \beta \in [\alpha_* - \delta, \alpha_* + \delta]$

$$\begin{aligned}
|\alpha^{k+1} - \alpha_*| &\stackrel{(6)}{\leq} \frac{-\varphi'''(\beta)}{2\varphi''(\alpha^k)} \cdot (\alpha^k - \alpha_*)^2 \\
&\leq \underbrace{\left[\frac{k_1}{2k_2}\right]}_{\leq 1} (\alpha^k - \alpha_*)^2
\end{aligned} \tag{26.0.1}$$

In general, $\left[\frac{k_1}{2k_2}\right]$ may be very large, but it is multiplied by $(\alpha^k - \alpha_*)^2$, which means that if we start close enough to α^* the upper-bound still makes sense.

$$|\alpha^{k+1} - \alpha_*| \leq \underbrace{\left[\frac{k_1}{2k_2}\right]}_{\leq 1} (\alpha^k - \alpha_*) \cdot (\alpha^k - \alpha_*)$$

hence $|\alpha^{k+1} - \alpha_*| < |\alpha^k - \alpha_*|$ and this proves the theorem. \square

This approach is pretty accurate, but it is usually undesirable to devote substantial resources to finding a value of α to precisely minimize f . This is because the computing resources needed to find a more precise minimum along one particular direction could instead be employed to identify a better search direction.

The following class of algorithms circumvent the problem of the existence of derivatives, but in general allow to find a solution without computing derivatives at all.

★ Mantra

The more derivatives we have, the smallest number of points we need (second derivative \rightarrow two points, third derivative \rightarrow zero points). The opposite holds as well.

Zero order algorithms

Definition 26.0.2 (Unimodal function). *Let $f : \mathbb{R} \rightarrow \mathbb{R}$. We say that f is unimodal if for some value $m \in \mathbb{R}$ it is monotonically increasing for each $x \leq m$ and monotonically decreasing for any $x \geq m$ or vice-versa. Notice that any unimodal function admits only one maximum (minimum) value $f(m)$.*

Let us be given φ , we are looking for α_* within an error tolerance ε . Notice that if the function φ is not unimodal we have no guarantee that the interval that we are discarding does not contain the “deepest” minimum. We know only the function values in given points and we would like to shrink a candidate interval in the best way possible.

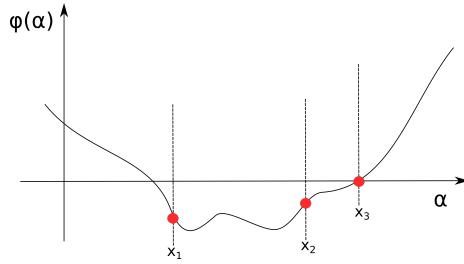


FIGURE 26.1: The candidate interval is $[x_1, x_2]$, since $\varphi(x_2) > \varphi(x_1)$ and we are allowed to exclude the interval $[x_3, +\infty)$ since the value in x_3 is bigger than $\varphi(x_2)$.

We propose an elegant solution via golden ratio.

Definition 26.0.3. *In mathematics, two quantities are in the **golden ratio** if their ratio is the same as the ratio of their sum to the larger of the two quantities. Formally, $\forall a, b \in \mathbb{R}$ such that $a > b > 0$,*

$$a + b : a = a : b$$

and we denote r_g such ratio.

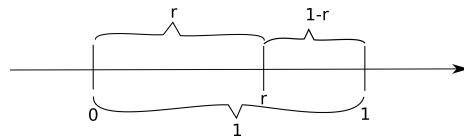


FIGURE 26.2: The relationship between r and $1 - r$ is $r : 1 = (1 - r) : r$.

Let us start from an interval $[\alpha_-, \alpha_+] \subseteq \mathbb{R}$, and let us take r as the inverse of the golden ratio $r = 1/\sqrt{5}-1/2 \approx 0.618$.

ALGORITHM 26.0.2 Pseudocode for zero-th order approach for local minimum detection.

```

1: procedure LSGRM( $\varphi, \alpha, \varepsilon$ )
2:    $\alpha_- \leftarrow 0;$ 
3:    $\alpha_+ \leftarrow \alpha;$ 
4:    $\alpha'_- \leftarrow (1 - r)\alpha;$ 
5:    $\alpha'_+ \leftarrow r\alpha;$ 
6:   while ( $\alpha_+ - \alpha_- > \varepsilon$ ) do ▷ not the same  $\varepsilon$ 
7:     if  $\varphi(v'_l) > \varphi(v'_r)$  then
8:        $\alpha_- \leftarrow \alpha'_-;$ 
9:        $\alpha'_- \leftarrow \alpha \leftarrow \alpha'_+;$ 
10:       $\alpha'_+ \leftarrow \alpha_+(1 - r)(\alpha_+ - \alpha_-);$ 
11:    else
12:       $\alpha_+ \leftarrow \alpha'_+;$ 
13:       $\alpha'_+ \leftarrow \alpha \leftarrow \alpha'_-;$ 
14:       $\alpha'_- \leftarrow \alpha_- + r(\alpha_+ - \alpha_-);$ 
15:    end if
16:   end while
17: end procedure

```

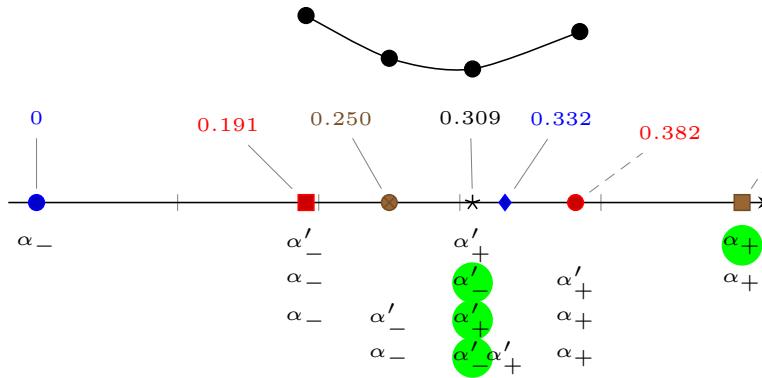


FIGURE 26.3: Plot of iterates of zero order algorithm for line search. The lines below the x axis represent the different values of α_- , α_+ , α'_- , α'_+ across different iterations, while the greed dot marks how the value α changes.

Example 26.0.1. Let us try Algorithm 26.0.2 on a toy example, displayed in Figure 26.3. Let $\alpha = 0$ and let $\varepsilon = 1$.

$$\begin{aligned}
\alpha_- &= 0 \\
\alpha_+ &= 0.5 \\
\alpha'_- &= r\alpha = (1 - 0.618) \cdot 0.5 = 0.191 \\
\alpha'_+ &= r \cdot \alpha = 0.618 \cdot 0.5 = 0.309
\end{aligned}$$

Since $0.5 - 0 < 1$ we enter the loop. According to the values of function φ in α_- and α_+ we enter the if branch:

$$\begin{aligned}
\alpha_- &= \alpha'_- = 0.191 \\
\alpha'_- &= \alpha'_+ = 0.309 \\
\alpha &= \alpha'_+ = 0.309 \\
\alpha'_+ &= \alpha_- + r \cdot (\alpha_+ - \alpha_-) = 0.191 + 0.618 \cdot (0.5 - 0.191) = 0.382
\end{aligned}$$

Provided that $0.5 - 0.191 < 1$ we enter the loop. According to the values of function φ in α_- and α_+ we enter the else branch:

$$\begin{aligned}
\alpha_+ &= \alpha'_+ = 0.382 \\
\alpha'_+ &= \alpha'_- = 0.309 \\
\alpha &= \alpha'_- = 0.309 \\
\alpha'_- &= \alpha_- + (1 - r) \cdot (\alpha_+ - \alpha_-) = 0.191 + 0.309 \cdot (0.382 - 0.191) = 0.25
\end{aligned}$$

Provided that $0.382 - 0.191 < 1$ we enter the loop. According to the values of function φ in α_- and α_+ we enter the if branch:

$$\begin{aligned}
\alpha_- &= \alpha'_- = 0.25 \\
\alpha'_- &= \alpha'_+ = 0.309 \\
\alpha &= \alpha'_+ = 0.309 \\
\alpha'_+ &= \alpha_- + r \cdot (\alpha_+ - \alpha_-) = 0.25 + 0.618 \cdot (0.382 - 0.25) = 0.332
\end{aligned}$$

Inexact line search

In the rest of this lecture we will present a line search method to determine the maximum amount to move along a given search direction, starting with a relatively large estimate of the step size for movement along the search direction, and iteratively shrinking the step size (“backtracking”) until a decrease of the objective function is observed that adequately corresponds to the decrease that is expected, based on the local gradient of the objective function.

Definition 26.0.4 (Armijo’s condition). *Let $\varphi : \mathbb{R}^+ \cup \{0\} \rightarrow \mathbb{R}$ such that $\varphi(\alpha) = f(\mathbf{x} + \alpha \mathbf{d})$. The **Armijo’s condition** selects those real values α such that the function value in those points ($\varphi(\alpha)$) is smaller than the value of a line which passes through 0 and is less steep than the tangent to the curve in such point.*

Formally, $\alpha \in \mathbb{R}$ satisfies **Armijo's condition** if for some control parameter $0 < m_1 < (\ll)1$ the following holds

$$\varphi(\alpha) \leq y_a(\alpha) = \varphi(0) + m_1 \alpha \varphi'(0) \quad (A)$$

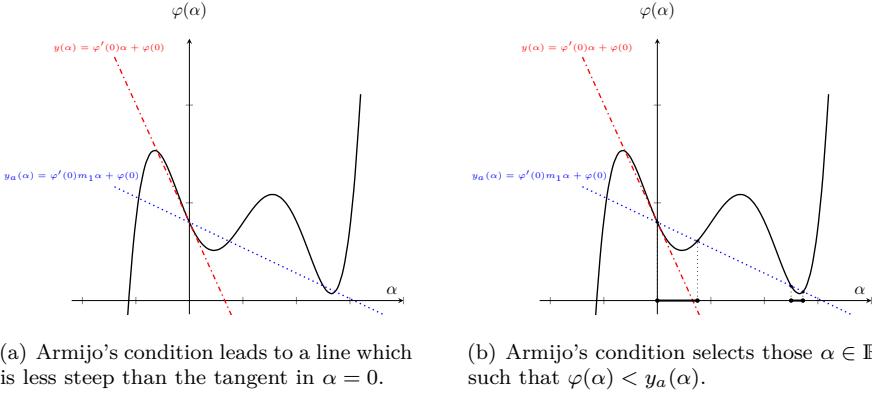


FIGURE 26.4: Graphical hint of the functioning of Armijo's condition.

Intuitively, provided that we are minimizing φ , $\varphi'(0) < 0$. This means that the slope of the line y_a ($m_1 \varphi'(0)$) is smaller in absolute value (hence bigger) than the slope of the tangent to the curve φ in 0, therefore the line y_a is less steep than y .

An attentive reader may notice that $\alpha \approx 0$ lead to values of $\varphi(\alpha) < y(\alpha)$, hence smaller than $y_a(\alpha)$, but this would lead to a very slow convergence. In order to overcome this issue one can decide to start from a large α and reduce it until it satisfies Armijo's condition (as shown in Algorithm 26.0.3) or introduce another algebraic constraint:

ALGORITHM 26.0.3 Pseudocode for backtracking line search.

```

1: procedure BLS( $\varphi, \varphi', \alpha, m_1, \tau$ )
2:   while ( $\varphi(\alpha) > \varphi(0) + m_1 \alpha \varphi'(0)$ ) do
3:      $\alpha \leftarrow \tau \alpha$ ; ▷ With  $\tau < 1$ 
4:   end while
5: end procedure

```

Definition 26.0.5 (Goldstein's condition). Let $\varphi : \mathbb{R}^+ \cup \{0\} \rightarrow \mathbb{R}$ such that $\varphi(\alpha) = f(\mathbf{x} + \alpha \mathbf{d})$ and let $m_1 \in \mathbb{R}$ be the Armijo's constant. The **Goldstein's condition** selects those real values α that satisfy Armijo's condition but also such that the function value in those points ($\varphi(\alpha)$) is bigger than the value of a line which passes through 0 and is less steep than the tangent to the curve in such point.

Formally, $\alpha \in \mathbb{R}$ satisfies the **Goldstein's condition** if for some control parameter $m_1 < m_2 < 1$ the following holds

$$\varphi(\alpha) \geq y_g(\alpha) = \varphi(0) + m_2\alpha\varphi'(0) \quad (G)$$

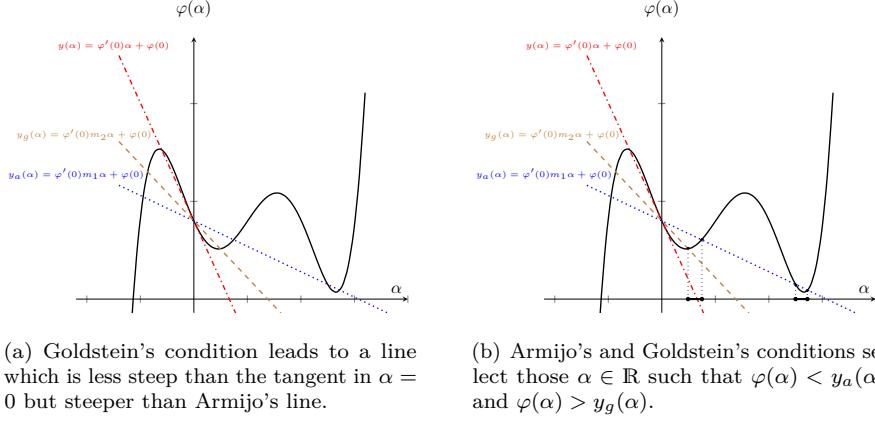


FIGURE 26.5: Conjunction of Armijo's and Goldstein's conditions.

It may happen that the points that satisfy both Goldstein's and Armijo's conditions does not contain a local minimum.

To circumvent this problem another condition comes to help us.

Definition 26.0.6 (Wolfe's condition). *Let $\varphi : \mathbb{R}^+ \cup \{0\} \rightarrow \mathbb{R}$ such that $\varphi(\alpha) = f(\mathbf{x} + \alpha\mathbf{d})$ and let $m_1 \in \mathbb{R}$ be the Armijo's constant. The **Wolfe's condition** selects those real values α that allow for a decrease in the slope (curvature condition), following the reasoning that close to the optimum the tangent line is almost horizontal.*

Formally, let $m_1 < m_3 < 1$

$$\varphi'(\alpha) \geq m_3\varphi'(0) \quad (W)$$

Figure 26.6(b) shows the intervals selected by Wolfe's rule. Provided that we are moving along a direction of negative curvature for f , we have that $\varphi'(\alpha) < 0$, therefore Wolfe's condition is satisfied also by those points α where the derivative is positive.

Sometimes, the curvature condition is modified to force the step length to stay in at least a broad neighborhood of a local minimizer or stationary point of the univariate function φ .

Provided that $\varphi'(0) < 0$, $|\varphi'(0)| = -\varphi'(0)$; follows a stronger variant of Wolfe's condition.

Definition 26.0.7 (Strong Wolfe's condition). *Let $\varphi : \mathbb{R}^+ \cup \{0\} \rightarrow \mathbb{R}$ such that $\varphi(\alpha) = f(\mathbf{x} + \alpha\mathbf{d})$ and let $m_1 \in \mathbb{R}$ be the Armijo's constant. The **strong***

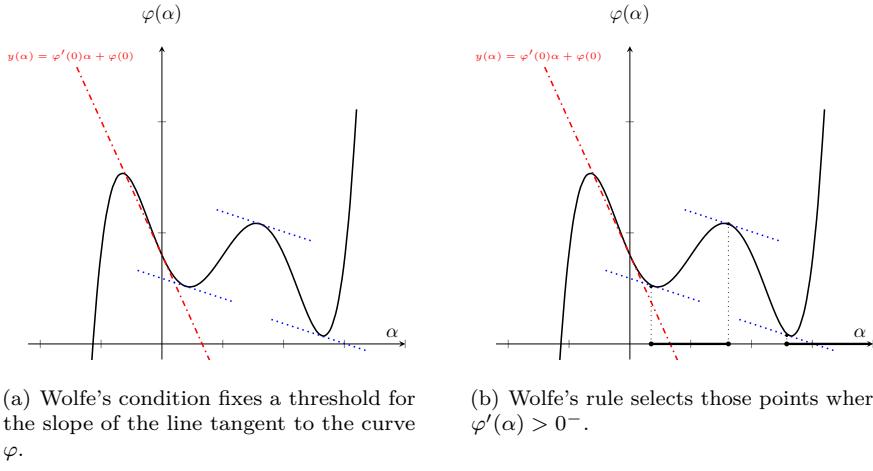


FIGURE 26.6: Wolfe's condition.

Wolfe's condition selects those real values α that allow for a decrease in the slope but prevent it to become big in absolute value. Formally, let $m_1 < m_3 < 1$

$$|\varphi'(\alpha)| \leq m_3 |\varphi'(0)| = -m_3 \varphi'(0) \quad (W')$$

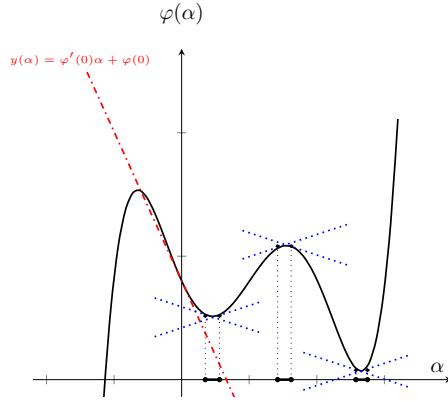


FIGURE 26.7: Strong Wolfe's condition.

Fact 26.0.3. Let $\varphi : \mathbb{R}^+ \cup \{0\} \rightarrow \mathbb{R}$ such that $\varphi(\alpha) = f(\mathbf{x} + \alpha \mathbf{d})$. If $\varphi'(\alpha) \geq 0$ and both Armijo's and Wolfe's (or Armijo's and strong Wolfe's) conditions hold then all local minima (maxima) are captured unless m_1 is too close to 1.

For a graphical hint of Proposition 26.0.3, see Figure 26.8, where we can

observe that if $m_1 \approx 1$, the line is pretty steep, therefore the intersection between Armijo's and Wolfe's regions is very small. In practice it is common to choose $m_1 \approx 0.0001$.

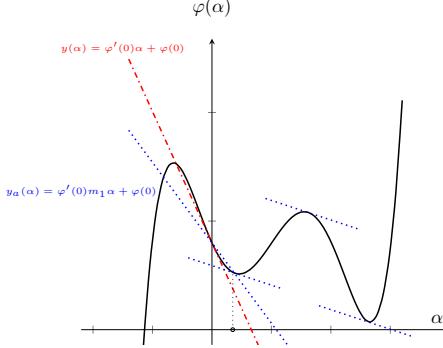


FIGURE 26.8: $m_1 \approx 1$ and the intersection between Armijo's and Wolfe's ranges.

The m_i are like the hyperparameters of machine learning. Less formally, if we choose an m_1 far enough from 1 everything works fine.

Theorem 26.0.4. Let $\varphi \in \mathcal{C}^1$ and $\varphi(\alpha)$ bounded below for $\alpha \geq 0$ then $\exists \alpha$ s.t. (A) \cap (W') holds.

Proof. Let us write Armijo's line: $y_a(\alpha) = \varphi(0) + m_1\varphi'(0)\alpha$. Let us define $d(\alpha) = y_a(\alpha) - \varphi(\alpha) = \varphi(0) + m_1\varphi'(0)\alpha - \varphi(\alpha)$, such that $d'(\alpha) = y'_a(\alpha) - \varphi'(\alpha) = m_1\varphi'(\alpha) - \varphi'(\alpha) = (m_1 - 1)\varphi'(\alpha)$. It goes without saying that $d(0) = 0$ and $d'(0) = \underbrace{(m_1 - 1)}_{<1} \underbrace{\varphi'(0)}_{<1} > 0$.

In $\alpha = 0$ the derivative of φ is negative, hence the function φ is decreasing. From the hypothesis we know that φ is bounded below, hence at a certain point it will start increasing its value until it will intersect with Armijo's line eventually. Let us call $\bar{\alpha} > 0$ the smallest value such that $d(\bar{\alpha}) = 0$. All the points $\alpha \in]0, \bar{\alpha}[$ satisfy Armijo's condition.

We are left with the task of proving that also strong Wolfe's condition holds for those $\alpha \in]0, \bar{\alpha}[$. 0 and $\bar{\alpha}$ are the two roots of the function d , so we can use Rolle's theorem, in order to prove that the function d has a stationary point in the interval $]0, \bar{\alpha}[$. Let us call the stationary point $\alpha^* \in]0, \bar{\alpha}[$, $d'(\alpha^*) = 0$ iff $\varphi'(\alpha^*) = -m_1\varphi'(0)$, therefore $|\varphi'(\alpha^*)| = m_1 \cdot |\varphi'(0)|$ and this proves Wolfe's rule, because $m_1 < m_3 < 1$, hence $|\varphi'(\alpha^*)| \leq m_3|\varphi'(0)|$. \square

How can we find such a point?

Fact 26.0.5. Let $\varphi : \mathbb{R}^+ \cup \{0\} \rightarrow \mathbb{R}$ such that $\varphi(\alpha) = f(\mathbf{x} + \alpha\mathbf{d})$ and let ∇f be L -Lipschitz. Then φ' is L -Lipschitz as well and L does not depend on x^i .

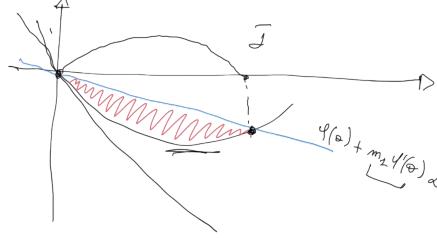


FIGURE 26.9: If φ is not going to $-\infty$ the Armijo's line (blue) and the function will meet in $\bar{\alpha} > 0$.

Proof.

$$\begin{aligned}
 |\varphi'(\alpha_1) - \varphi'(\alpha_2)| &= |\mathbf{d}^T \nabla f(\mathbf{x} + \alpha_1 \mathbf{d}) - \mathbf{d}^T \nabla f(\mathbf{x} + \alpha_2 \mathbf{d})| \\
 &= \mathbf{d}^T \underbrace{|\nabla f(\mathbf{x} + \alpha_1 \mathbf{d}) - \nabla f(\mathbf{x} + \alpha_2 \mathbf{d})|}_{\leq L \cdot |\mathbf{x} + \alpha_1 \mathbf{d} - (\mathbf{x} + \alpha_2 \mathbf{d})|} \\
 &\leq \mathbf{d}^T L \cdot |\mathbf{x} + \alpha_1 \mathbf{d} - (\mathbf{x} + \alpha_2 \mathbf{d})| \\
 &= \mathbf{d}^T L |\alpha_1 - \alpha_2| \mathbf{d} \\
 &\stackrel{*}{=} L \cdot |\alpha_1 - \alpha_2|
 \end{aligned}$$

where $\stackrel{*}{=}$ follows from the fact that \mathbf{d} has norm 1. \square

Fact 26.0.6. Let $\varphi : \mathbb{R}^+ \cup \{0\} \rightarrow \mathbb{R}$ such that $\varphi(\alpha) = f(\mathbf{x} + \alpha \mathbf{d})$, φ bounded below and ∇f L -Lipschitz. $\bar{\alpha} \in \mathbb{R}$ has the same behaviour of $\|\nabla f(\mathbf{x}^i)\|$. Formally,

$$\bar{\alpha} > (1 - m_1) \frac{\|\nabla f(\mathbf{x}^i)\|}{L}$$

Proof. According to Proposition 26.0.5, φ' is L -Lipschitz and

$$L(\bar{\alpha} - 0) \geq \varphi'(\bar{\alpha}) - \varphi'(0) > (1 - m_1)(-\varphi'(0)) \stackrel{*}{=} (1 - m_1)(\|\nabla f(\mathbf{x}^i)\|)$$

where $\stackrel{*}{=}$ follows from

$$\varphi'(\alpha) = \langle \nabla f(\mathbf{x} + \alpha \mathbf{d}), \mathbf{d} \rangle = \frac{-\nabla f(\mathbf{x} + \alpha \mathbf{d}) / \|\nabla f(\mathbf{x} + \alpha \mathbf{d})\|}{\|\nabla f(\mathbf{x} + \alpha \mathbf{d})\|^2}$$

$$\text{therefore } \bar{\alpha} > (1 - m_1) \frac{\|\nabla f(\mathbf{x}^i)\|}{L}.$$

\square

Now we can prove the following

Theorem 26.0.7. If (A) \cap (W) holds $\forall i$ then either $\{f(\mathbf{x}^i)\} \rightarrow -\infty$ or $\{\|\nabla f(\mathbf{x}^i)\|\} \rightarrow 0$.

Proof. By contradiction, we assume that $\|\nabla f(\mathbf{x}^i)\|$ diverges. Formally, $-\varphi'(0) = \|\nabla f(\mathbf{x}^i)\| \geq \varepsilon > 0 \forall i$. Then

1. (W) $\implies \alpha^i \geq \bar{\alpha} > (1 - m_1) \frac{\|\nabla f(\mathbf{x}^i)\|}{L}$. Let $\delta := \frac{(1-m_1)\cdot\varepsilon}{L}$, then $\alpha^i \geq \delta > 0$;
2. (A) $\implies f(\mathbf{x}^{i+1}) \leq f(\mathbf{x}^i) - m_1 \alpha^i \|\nabla f(\mathbf{x}^i)\| \leq f(\mathbf{x}^i) - m_1 \delta \varepsilon$;
3. So $\{f(\mathbf{x}^i)\} \rightarrow -\infty$ (or $\{\|\nabla f(\mathbf{x}^i)\|\} \rightarrow 0$).

□

Backtracking is similar: for simplicity, $\alpha = 1$ (input)
 $\|\nabla f(\mathbf{x}^i)\| > \varepsilon \forall i \implies \bar{\alpha} > \delta > 0 \forall i$
 $h = \min\{k : \tau^{-k} \leq \delta\} \implies \alpha^i \geq \tau^{-h} > 0 \forall i \implies f(\mathbf{x}^{i+1}) \leq f(\mathbf{x}^i) - m_1 \tau^{-h} \varepsilon \implies \{f(\mathbf{x}^i)\} \rightarrow -\infty$ or ↘.

Lecture 27

8th of November 2018

A. Frangioni

27.1 Good practices for the project's design

In this lecture some good suggestions for implementing the gradient method in Matlab:

- When we have to implement a function the first question we need to ask ourselves is: “how many parameters should the function take?”;
- The interface needs to be designed, i.e. in the case of the gradient method we would like the user to be able to run it although he might not know what a good starting point could be;
- There might be some parameters with the same semantic of hyper-parameters, so they need to be adjusted in order to specify the algorithm;
- Another thing that should be taken into consideration is that sometimes, when measuring the number of times the code of the function has been executed we need to count also the calls to functions inside a single step;
- It is important to check the conditions that should be satisfied by the input data for the theoretical coherence. In case of not valid input an error message should be returned;
- Another important thing to be done is building a set of test functions, possibly limit case for the problem, to see how the algorithm works;
- `diff(f, x)` calculates the gradient of the function f in the variable x . If we want to get a simplified version, we may run `simplify(diff(f,x))`;
- A very nice tool to compute derivatives for arbitrary functions is **Adigator***

*<https://sourceforge.net/projects/adigator>

Lecture 28

14th of November 2018

A. Frangioni

★ Mantra

If you want better convergence, use a better model.

28.1 General descent methods

💡 Do you recall?

So far, we chose as direction for the step the direction where the decrease of the function is maximum ($\mathbf{d}^i = -\nabla f(\mathbf{x}^i)$) and for that descent direction we applied a line search for finding the optimal step size α . This was possible thanks to the **convergence argument** which says that in proximity of the stationary point of the linear model the norm of the gradient goes to 0. Formally, $\|\nabla f(\mathbf{x})\| \rightarrow 0$ when $\varphi'(0) \rightarrow 0$, where $\varphi'(\alpha) = <\nabla f(\mathbf{x}^i + \alpha^i \mathbf{d}^i), \mathbf{d}^i>$.

In this lecture we will propose a different choice for the moving direction \mathbf{d}^i , that keeps satisfying the convergence argument.

For example, let us take as direction a rotation of the opposite of the gradient, the value of φ' is then the cosine of the angle of the rotation times the opposite of the norm of the gradient. We have an infinite number of angles that we can choose, so we have a lot of flexibility.

Let us define the angle θ^i between the search direction \mathbf{d}^i and the steepest descent direction $-\nabla f(\mathbf{x}^i)$ as

$$\cos \theta^i = \frac{\langle \nabla f(\mathbf{x}), \mathbf{d}^i \rangle}{\|\nabla f(\mathbf{x}^i)\| \cdot \|\mathbf{d}^i\|}$$

Note that the angle between \mathbf{d}^i and the gradient should not be too close to 90° , otherwise the cosine would get approximately 0..

Theorem 28.1.1 (Zoutendijk's theorem). *Let us consider minimum problem (P) where the objective function $f \in \mathcal{C}^1$, ∇f is L-Lipschitz and f bounded below. Let us consider an iteration $\mathbf{x}^{i+1} = \mathbf{x}^i + \alpha^i \mathbf{d}^i$, where \mathbf{d}^i is a descent direction and $\alpha^i \in \mathbb{R}$ satisfies both Armijo's and Wolfe's conditions. Then*

$$\sum_{i=1}^{\infty} \cos^2(\theta^i) \|\nabla f(\mathbf{x}^i)\|^2 < \infty$$

 Do you recall?

A positive infinite sequence, whose corresponding series converges to a number, converges to 0 reasonably fast.

Therefore, if we choose an angle θ^i which is bounded below then the norm of the gradient goes to 0 quickly. More formally,

Fact 28.1.2. *Under the same hypotheses of Theorem 28.1.1, $\cos(\theta^i) \geq \varepsilon > 0 \implies \|\nabla f(\mathbf{x}^i)\| \rightarrow 0$.*

Proof. As stated in the recap, the following holds:

$$\cos^2(\theta^i) \|\nabla f(\mathbf{x}^i)\|^2 \rightarrow 0$$

therefore, one between $\cos(\theta^i)$ and $\|\nabla f(\mathbf{x}^i)\|^2$ should converge to zero, but it is no the cosine, because it is bounded below by $\varepsilon > 0$. \square

28.2 Newton's method

Newton's method attempts to solve the minimum problem (P) by constructing a sequence $\{\mathbf{x}^i\}$ from an initial guess (starting point) $\mathbf{x}^0 \in \mathbb{R}^n$ that converges towards a minimizer \mathbf{x}_* of f by using a sequence of second-order Taylor approximations of f around the iterates.

28.2.1 Convex case

In the next line we will explore Newton's method in the simplest case, when the function is perfectly convex (i.e. the Hessian is strictly positive definite).

Theorem 28.2.1. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ be a function such that $\nabla^2 f(\mathbf{x}^i) \succ 0$. Then the second order model in proximity of point \mathbf{x} ($Q_{\mathbf{x}^i}(\mathbf{y})$ for $\mathbf{y} \in \mathcal{B}(\mathbf{x}, \delta)$) admits a minimum.

Proof. Let us write \mathbf{x} instead of \mathbf{x}^i to ease notation. The second order Taylor's expansion of f around the point \mathbf{x} (recall Definition 20.0.7) is:

$$\begin{aligned} Q_{\mathbf{x}}(\mathbf{y}) &= f(\mathbf{x}) + \langle \nabla f(\mathbf{x}), \mathbf{y} - \mathbf{x} \rangle + \frac{1}{2} (\mathbf{y} - \mathbf{x})^T \nabla^2 f(\mathbf{x}) (\mathbf{y} - \mathbf{x}) \\ &= f(\mathbf{x}) + \langle \nabla f(\mathbf{x}), \mathbf{y} \rangle - \langle \nabla f(\mathbf{x}), \mathbf{x} \rangle + \\ &\quad + \frac{1}{2} \left(\mathbf{y}^T \nabla^2 f(\mathbf{x}) \mathbf{y} - 2\mathbf{x}^T \nabla^2 f(\mathbf{x}) \mathbf{y} + \mathbf{x}^T \nabla^2 f(\mathbf{x}) \mathbf{x} \right) \\ &= f(\mathbf{x}) + \langle \nabla f(\mathbf{x}), \mathbf{y} \rangle - \langle \nabla f(\mathbf{x}), \mathbf{x} \rangle + \\ &\quad + \frac{1}{2} \mathbf{y}^T \nabla^2 f(\mathbf{x}) \mathbf{y} - \mathbf{x}^T \nabla^2 f(\mathbf{x}) \mathbf{y} + \frac{1}{2} \mathbf{x}^T \nabla^2 f(\mathbf{x}) \mathbf{x} \end{aligned}$$

We are looking for the minimum of $Q_{\mathbf{x}}$, that is a stationary point. Let us differentiate $Q_{\mathbf{x}}$:

$$\begin{aligned} \frac{\partial Q_{\mathbf{x}}(\mathbf{y})}{\partial \mathbf{y}} &\stackrel{*}{=} \frac{\partial (\langle \nabla f(\mathbf{x}), \mathbf{y} \rangle + \frac{1}{2} \mathbf{y}^T \nabla^2 f(\mathbf{x}) \mathbf{y} - \mathbf{x}^T \nabla^2 f(\mathbf{x}) \mathbf{y})}{\partial \mathbf{y}} \\ &= \frac{\partial (\langle \nabla f(\mathbf{x}) + \frac{1}{2} \mathbf{y}^T \nabla^2 f(\mathbf{x}) - \mathbf{x}^T \nabla^2 f(\mathbf{x}), \mathbf{y} \rangle)}{\partial \mathbf{y}} \\ &= \nabla f(\mathbf{x}) - \mathbf{x}^T \nabla^2 f(\mathbf{x}) + \frac{1}{2} \nabla^2 f(\mathbf{x}) \mathbf{y} \end{aligned}$$

where $\stackrel{*}{=}$ is given by the fact that $f(\mathbf{x})$, $-\langle \nabla f(\mathbf{x}), \mathbf{x} \rangle$ and $\frac{1}{2} \mathbf{x}^T \nabla^2 f(\mathbf{x}) \mathbf{x}$ are constant terms with respect to \mathbf{y} . The equality obtained above allows to say that \mathbf{y} is a stationary point iff

$$\begin{aligned} \frac{1}{2} \nabla^2 f(\mathbf{x}) \mathbf{y} &= \nabla f^2(\mathbf{x}) \mathbf{x} - \nabla f(\mathbf{x}) \\ \mathbf{y} &= 2 \cdot \left(\mathbf{x} - [\nabla^2 f(\mathbf{x})]^{-1} \nabla f(\mathbf{x}) \right) \end{aligned}$$

□

The moving direction of the gradient descent algorithm using Newton's method is obtained taking the opposite of the inverse of the Hessian times the gradient of the function. Formally, $\mathbf{d}^i = -[\nabla^2 f(\mathbf{x}^i)]^{-1} \nabla f(\mathbf{x}^i)$.

An attentive reader may notice that Theorem 28.2.1 holds only if the Hessian is invertible.

Moreover, even if the Hessian is invertible, the computation of the descent direction requires to solve a linear system. A way to circumvent this problem is putting the gradient to 0, so we can write the Taylor form of the gradient and solve a linear equation which is $\nabla f(\mathbf{x}) \approx \nabla f(\mathbf{x}^i) + \nabla^2 f(\mathbf{x}^i)(\mathbf{x} - \mathbf{x}^i)$.

Theorem 28.2.2. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ s.t. $f \in \mathcal{C}^3$ and let \mathbf{x}_* be a stationary point such that $\nabla f(\mathbf{x}_*) = 0$ and $\nabla^2 f(\mathbf{x}_*) \succ 0$. Then $\exists \mathcal{B}(\mathbf{x}_*, r)$ s.t. $\mathbf{x}^1 \in \mathcal{B}$ and $\{\mathbf{x}^i\} \rightarrow \mathbf{x}_*$ quadratically.

Notice that Theorem 28.2.2 states that the descent direction computed through Newton's method is correct both close and far from the minimum point.

An attentive reader may notice that the scalar product between the function value at the i -th iterate $f(\mathbf{x}^i)$ and the descent direction \mathbf{d}^i should be negative but also not too close to 0. Such condition is ensured when the function f is such that $uI \preceq \nabla^2 f \preceq LI$, which implies that the function is strongly convex, in other words that the eigenvalues of the Hessian do not get too close to zero (in terms of machine precision).

Theorem 28.2.3 (Global convergence of Newton's method). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ s.t. $f \in \mathcal{C}^3$ and such that satisfies $uI \preceq \nabla^2 f \preceq LI$ and $\cos \theta^i = \frac{\langle \nabla f(\mathbf{x}^i), \mathbf{d}^i \rangle}{\|\nabla f(\mathbf{x}^i)\| \cdot \|\mathbf{d}^i\|}$. Then Newton's method converges globally.

Proof.

Aggiungere la ref alla variational characterization nella parte di Numerical Methods

The variational characterization of the eigenvalues and the definition of \mathbf{d}^i ($\mathbf{d}^i := -[\nabla^2 f(\mathbf{x}^i)]^{-1} \nabla f(\mathbf{x}^i) \iff \nabla^2 f(\mathbf{x}^i) \mathbf{d}^i = -\nabla f(\mathbf{x}^i)$) leads to the following upper-bounds:

1.

$$\mathbf{d}^{i^T} \nabla f(\mathbf{x}^i) = -(\mathbf{d}^i)^T \nabla^2 f(\mathbf{x}^i) \mathbf{d}^i \leq -\lambda_{\min} \|\mathbf{d}^i\|^2$$

2.

$$\|\nabla f(\mathbf{x}^i)\| = \|\nabla^2 f(\mathbf{x}^i) \mathbf{d}^i\| \leq \|\nabla^2 f(\mathbf{x}^i)\| \cdot \|\mathbf{d}^i\| = \lambda_{\max} \|\mathbf{d}^i\|$$

From the definition of $\cos \theta^i = \frac{\langle \nabla f(\mathbf{x}^i), \mathbf{d}^i \rangle}{\|\nabla f(\mathbf{x}^i)\| \cdot \|\mathbf{d}^i\|}$ we get

$$\cos(\theta^i) \leq -\lambda_{\min}/\lambda_{\max} \leq -u/L$$

□

Theorem 28.2.4. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ s.t. $f \in \mathcal{C}^3$ and such that satisfies $uI \preceq \nabla^2 f \preceq LI$ and $\cos(\theta^i) \leq -\lambda_{\min}/\lambda_{\max} \leq -u/L$. Then the method not only converges, but we also have that from some iteration onward, $\alpha^i = 1$ always satisfies Armijo's condition.

Proof.

$$\begin{aligned}
f(\mathbf{x}^i + \mathbf{d}^i) &= f(\mathbf{x}^i) + \langle \nabla f(\mathbf{x}^i), \mathbf{d}^i \rangle + \frac{1}{2} \cdot (\mathbf{d}^i)^T [\nabla^2 f(\mathbf{x}^i)] \mathbf{d}^i + R_3(\|\mathbf{d}^i\|) \\
&= f(\mathbf{x}^i) + (\mathbf{d}^i)^T \nabla f(\mathbf{x}^i) + \frac{1}{2} \cdot (\mathbf{d}^i)^T [\nabla^2 f(\mathbf{x}^i)] \mathbf{d}^i + R_3(\|\mathbf{d}^i\|) \\
&= f(\mathbf{x}^i) - \nabla f(\mathbf{x}^i)^T \left([\nabla^2 f(\mathbf{x}^i)]^{-1} \right)^T \nabla f(\mathbf{x}^i) + \\
&\quad + \frac{1}{2} \nabla f(\mathbf{x}^i)^T \left([\nabla^2 f(\mathbf{x}^i)]^{-1} \right)^T \cancel{\nabla^2 f(\mathbf{x}^i)} \cancel{[\nabla^2 f(\mathbf{x}^i)]^{-1}} \nabla f(\mathbf{x}^i) + R_3(\|\mathbf{d}^i\|) \\
&\stackrel{*}{=} f(\mathbf{x}^i) - \nabla f(\mathbf{x}^i)^T [\nabla^2 f(\mathbf{x}^i)]^{-1} \nabla f(\mathbf{x}^i) + \\
&\quad + \frac{1}{2} \nabla f(\mathbf{x}^i)^T [\nabla^2 f(\mathbf{x}^i)]^{-1} \nabla f(\mathbf{x}^i) + R_3(\|\mathbf{d}^i\|) \\
&= f(\mathbf{x}^i) - \frac{1}{2} \cdot \nabla f(\mathbf{x}^i)^T \cdot \underbrace{[\nabla^2 f(\mathbf{x}^i)]^{-1} \nabla f(\mathbf{x}^i)}_{\mathbf{d}^i} + R_3(\|\mathbf{d}^i\|) \\
&= f(\mathbf{x}^i) - \frac{1}{2} \langle \nabla f(\mathbf{x}^i), \mathbf{d}^i \rangle + R_3(\|\mathbf{d}^i\|)
\end{aligned} \tag{28.2.1}$$

where $\stackrel{*}{=}$ follows from Corollary 20.0.3 ($f \in \mathcal{C}^2 \Rightarrow \nabla^2 f \in S(n, \mathbb{R})$) and the fact that the inverse of a symmetric matrix (if it exists) is a symmetric matrix as well. \square

It can be proved that the convergence is superlinear.

If we start with a step size of 1, we end up in a situation in which the line search isn't computed when we are close to the minimum.

This works under the assumption that the eigenvalues are bounded both above and below (deriving from the bounds on the Hessian).

28.2.2 Interpretation of Newton's method

Riscrivere meglio

Let us consider Newton's method from a different point of view. We can construct a different space where the gradient method coincides with Newton's method. Given R which is not singular, we make a variable change ($\mathbf{y} = R\mathbf{x}$) — which is possible given that R is non singular — and we get $f_{\mathbf{y}}(\mathbf{y}) = \frac{1}{2}\mathbf{y}^T I\mathbf{y} + qR^{-1}\mathbf{y}$, which has as an Hessian the identity matrix, which is the optimal matrix for convergence in Newton's method.

Formally, the descending direction $\mathbf{d}_{\mathbf{y}}$ is computed as follows: $\mathbf{d}_{\mathbf{y}} = -\nabla f_{\mathbf{y}}(\mathbf{y}) = -\mathbf{y} - R^{-1}q$. Since we chose 1 as step size we obtain that $\nabla f_{\mathbf{y}}(\mathbf{y} + \mathbf{d}_{\mathbf{y}}) = \nabla f_{\mathbf{y}}(\mathbf{y} - \mathbf{y} - R^{-1}\mathbf{q}) = \nabla f_{\mathbf{y}}(-R^{-1}\mathbf{q}) = 0$.

It takes only one iteration, because all the eigenvalues are 1 so the the ratio between the greatest and the smallest is 1 and the subtraction is 0.

If we do the inverse operation ($\mathbf{x} = R^{-1}\mathbf{y}$) to the direction we get the direction in \mathbf{x} variable.

Problem:

We made a lot of assumptions on the Hessian, without the ones there's no guarantee that we are moving on a descending direction. How can we relax these constraints?

28.2.3 Non convex case

Saying that the function f is not convex is equivalent to say that the Hessian is indefinite. In such case we propose to compute the descent direction (previously defined as $\mathbf{d}^i = -[\nabla^2 f(\mathbf{x}^i)]^{-1} \nabla f(\mathbf{x}^i)$) substituting the inverse of the Hessian with a matrix H' which is strictly positive definite ad has more or less the same properties of the Hessian.

For example, we can build H^i summing to the Hessian a multiple of the identity matrix: $H^i = \nabla^2 f(\mathbf{x}^i) + \varepsilon^i I \succ 0$. The sequence $\{H^i\}$ is supposed to converge to a matrix H' that is strictly positive definite.

A proper choice for ε should ensure that all the eigenvalues are at least δ , for an “appropriately chosen smallish δ ”. Formally,

$$\varepsilon = \max\{0, \delta - \lambda_{\min}\}$$

The real value δ should not be too small for both numerical (any double $\leq 1e-16$ is treated as 0 by computers) and algorithmic (the smaller $\lambda_{\min}(\nabla^2 f(\mathbf{x}^i) + \varepsilon I)$ the more elongated the axes of $S(Q_{\mathbf{x}^i}, \cdot)$) reasons.

It can be proved that the ε we chose is the solution to an optimization problem: $\min\{\|H - \nabla^2 f(\mathbf{x}^i)\| \mid H \succeq \delta I\}$. The choice for δ in the Matlab code is 10^{-6} .

Observation 28.2.1. Note that these constraints are important and we will get back to them later on in the course.

Could we use a different norm instead of the 2-norm? Yes, for example we can use Frobenius's norm, changing a bit the algorithm, but we still get convergence. We would need to solve $\min\{\|H - \nabla^2 f(\mathbf{x}^i)\|_F \mid H \succeq \delta I\}$, which is performed in two steps:

1. compute spectral decomposition $\nabla^2 f(\mathbf{x}^i) = H \Lambda H^T$
2. $H^i = H \bar{\Lambda} H^T$ with $\bar{\gamma}^i = \max\{\lambda^i, \delta\}$

In both cases, $\{\mathbf{x}^i\} \rightarrow \mathbf{x}_*$ with $\nabla^2 f(\mathbf{x}_*) \succeq \delta I$, which implies $\varepsilon^i = 0$ and $H^i = \nabla^2 f(\mathbf{x}^i)$ eventually. It holds also that we have superlinear convergence if “ H^i looks like $\nabla^2 f(\mathbf{x}^i)$ along \mathbf{d}^i ”, formally $\lim_{i \rightarrow \infty} \|(H^i - \nabla^2 f(\mathbf{x}^i))\mathbf{d}^i\| \|\mathbf{d}^i\| = 0$.

Computational complexity

We still need to compute eigenvalues, which takes $O(n^3)$, which is too much if we are in multidimensional spaces.

As a closing observation we want to stress that Newton's method is very fast to converge to a local minimum, but this behaviour may represent a problem, because the algorithm may miss global minima.

Lecture 29

16th of November 2018

A. Frangioni

What happens when the Hessian isn't strictly positive definite? If there are some negative eigenvalues, I can desume that there are some directions in which the function goes to $-\infty$. The model has no minimum, unless we restrict to a **compact set**.

In particular, we may decide to restrict to a part of the space where we can trust our model, which is called **trust region**.

Finding such a region is a NP-hard problem, if we don't restrict to a ball.

Definition 29.0.1 (Karush-Khun-Tucker conditions). *Any optimal solution of the problem \mathbf{x}^{i+1} must satisfy that $\exists \lambda \geq 0$ s.t.*

KARUSH: $[H^i + \lambda I]\mathbf{x}^{i+1} = -\nabla f(\mathbf{x}^i)$ [linear];

KUHN: $H^i + \lambda I \succeq 0$ [semidefinite];

TUCKER: $\lambda(r - \|\mathbf{x}^{i+1}\|) = 0$ [nonlinear].

Where the last condition means that λ has to be 0, unless the solution we get is exactly on the border of the ball.

What's the difference between this approach and what we used to do before?
We have two different cases:

- $\|\mathbf{x}^{i+1}\| < r \implies \lambda = 0 \implies$ normal Newton step (\mathcal{T} has no effect);
- $\lambda > 0 (= \text{small radius } r) \implies$ like in line search with $\varepsilon^i = \lambda$.

The problem is computing these systems of equations taking less than $O(n^3)$.

Key idea

We don't need to compute the Hessian, we can use the first order information to infer things on the second order matrix, although we don't really need to compute the Hessian.

29.1 Quasi-newton methods

At a step we have: $m^i(x) = \nabla f(\mathbf{x}^i)(x - \mathbf{x}^i) + \frac{1}{2}(x - \mathbf{x}^i)^T H^i(x - \mathbf{x}^i)$, $\mathbf{x}^{i+1} = \mathbf{x}^i + \alpha^i d^i$

At the next step we recompute the gradient in \mathbf{x}^{i+1} and the matrix H^{i+1} :
 $m^{i+1}(x) = \nabla f(\mathbf{x}^{i+1})(x - \mathbf{x}^{i+1}) + \frac{1}{2}(x - \mathbf{x}^{i+1})^T H^{i+1}(x - \mathbf{x}^{i+1})$

How should H^i be chosen? It should satisfy the following:

1. positive definite ($H^{i+1} \succ 0$);
2. we know the gradient in the previous point, we know the gradient in the current point, we construct H^{i+1} such that the model works: $\nabla m^{i+1}(\mathbf{x}^i) = \nabla f(\mathbf{x}^i)$;
3. $\|H^{i+1} - H^i\|$ is “small”.

This new model isn't too different from the previous one, because of the third property. Or equivalently $H^{i+1}(\mathbf{x}^{i+1} - \mathbf{x}^i) = \nabla f(\mathbf{x}^{i+1}) - \nabla f(\mathbf{x}^i)$, which we call **secant equation** and we denote (S).

To ease notation we define s^i such that: $s^i = \mathbf{x}^{i+1} - \mathbf{x}^i = \alpha^i d^i$ and $y^i = \nabla f(\mathbf{x}^{i+1}) - \nabla f(\mathbf{x}^i)$.

s^i is chosen, while y^i is decided by the function.

In order to have a matrix H^i that satisfies the first two condition we could check that $H^{i+1}s^i = y^i$, because this implies $s^i y^i = (s^i)^T H^{i+1} s^i$ and this implies 1. and 2., hence we obtain the **curvature condition** $s^i y^i > 0$.

Theorem 29.1.1. *Wolf condition implies $s^i y^i > 0$, using the notation we introduced: (W) \implies (C).*

Proof.

$$\begin{aligned} \varphi'(\alpha^i) &= \nabla f(\mathbf{x}^{i+1})d^i \geq m_3 \varphi'(0) = m_3 \nabla f(\mathbf{x}^i)d^i \\ &\Downarrow \\ (\nabla f(\mathbf{x}^{i+1}) - \nabla f(\mathbf{x}^i))d^i &\geq (m_3 - 1)\varphi'(0) > 0 \end{aligned}$$

□

Observation 29.1.1. *We may observe that this theorem implies that if we perform Armijo Wolf exact line search condition (C) can always be satisfied.*

29.1.1 Davidson-Fletcher-Powell

How can we choose a H^i that satisfies the three conditions enumerated above? Taking $H^{i+1} = \operatorname{argmin}\{\|H - H^i\| : H \in (S) \text{ and } H \succeq 0\}$ is a good idea and for this minimum problem holds the following:

Theorem 29.1.2 (Davidson-Feltcher-Powell). *The new matrix is obtained at each step constructing a rank two matrix, obtained from H^i as a rank two correction, as follows: $H^{i+1} = (I - \rho^i y^i (s^i)^T) H^i (I - \rho^i s^i (y^i)^T) + \rho^i y^i (y^i)^T$*

Let us denote $B^i = H^{i-1}$. At any step we need to compute $B^{i+1} = (H^{i+1})^{-1}$, because we need to solve the system. We have some formulas that give us a way to compute $(H^{i+1})^{-1}$ from H^{i-1} .

Theorem 29.1.3 (Sherman-Morrison-Woodbury). *The inverse of a matrix of the form $A + ab^T$ has the following shape: $[A + ab^T]^{-1} = \frac{A^{-1} - A^{-1}ab^TA^{-1}}{1 - b^TA^{-1}a}$.*

Observation 29.1.2. *From Theorem 29.1.3 we can conclude that $B^{i+1} = \frac{B^i + \rho^i s^i (s^i)^T - B^i y^i (y^i)^T B^i}{(y^i)^T B^i y^i}$.*

It's important to notice that this operation has a cost of $O(n^2)$.

We can do better, in terms of computational complexity.

29.1.2 Broyden-Fletcher-Goldfarb-Shanno

We can use directly B^i , the inverse of H^i . Write (S) for B^{i+1} : $s^i = B^{i+1}y^i \implies B^{i+1} = \arg\min \{\|B - B^i\| : \dots\}$.

$$B^{i+1} = B^i + \rho^i [(1 + \rho^i (y^i)^T B^i y^i) s^i (s^i)^T - (B^i y^i (s^i)^T + s^i (y^i)^T B^i)]$$

This formula proves to be more stable than the other one.

This method takes $O(n^2)$.

The two B^i 's, obtained from DFP and BFGS, are different although both sensible, but we can use a convex combination of the two.

Observation 29.1.3. *How can we choose H^1 ? The value we choose will make a difference in the results, at least for the first steps.*

Let us see a couple of choices for B^1 :

- Scalar multiples of identity, but how to choose the scalar?
- Compute the gradient in every direction and approximate H . This will cost $O(n^3)$, but it should be done only once.

Let us compute the space needed to store the B^i 's: order of n^2 is still a lot. What happens if we restrict to working with information of the last k operations?

29.1.3 Poorman's approach

At each step we only consider B^{i-k} and k rank one operations. This operations cost n each, and we have k lines. The problem is that B^{i-k} takes $O(n^2)$ space. We can optimize if we choose B^{i-k} to be simpler, say a multiple of the identity, or finite difference of the gradient. Then the space complexity is $O(kn)$.

I need to tune the algorithm to find the right k which gives me enough precision and also keeps the computational cost low.

Final observation of quasi Newton methods

We may notice that this variation of Newton method doesn't get trapped in local minima, as Newton method did. In the end, the fact that quasi Newton isn't that precise at the beginning may be a good feature.

29.2 Conjugate gradient method

💡 Do you recall?

In the gradient method, the angle between two consecutive directions is exactly 90° , as can be seen in Figure 29.1.

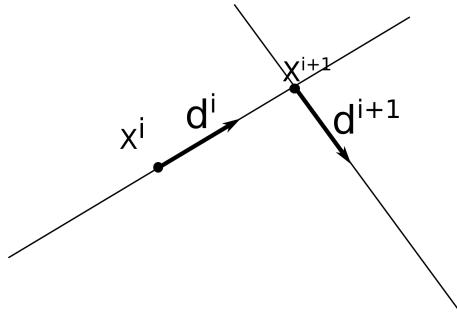


FIGURE 29.1: Geometric idea on how the new direction is chosen.

We would like to take into account not only the subspace spanned by d^{i+1} but we would like to optimize over larger and larger subspaces (spanned by d^i and d^{i+1}).

Definition 29.2.1 (Q-conjugate). Let v and w be vectors in \mathbb{R} . We say that v and w are **Q-conjugate** if $(v)^T Q w = 0$.

We would like pick a direction to be Q -conjugate with all the previous iterations. The point is that we can't take into account all the previous directions, but we will see that we only need the previous direction to obtain all the information we need.

ALGORITHM 29.2.1 Pseudocode for conjugate gradient method for quadratic functions.

```

1: procedure CGQ( $Q, q, x, \varepsilon$ )
2:    $d^- \leftarrow 0;$ 
3:   while ( $\|\nabla f(x)\| > \varepsilon$ ) do
4:     if ( $d^- = 0$ ) then
5:        $d \leftarrow -\nabla f(x);$ 
6:     else
7:        $\beta = (\nabla f(x)^T Q d^-) / ((d^-)^T Q d^-);$ 
8:        $d \leftarrow -\nabla f(x) + \beta d^-;$ 
9:     end if
10:     $\alpha \leftarrow (\nabla f(x)^T d) / (d^T Q d);$ 
11:     $x \leftarrow x + \alpha d;$ 
12:     $d^- \leftarrow d;$ 
13:   end while
14: end procedure

```

The number of iterations needed to converge is proportional to the clusterization of the eigenvalues of the matrix Q .

The algorithm that was presented is for quadratic functions, but the same algorithm works for non quadratic function as well, as long as we change the formula for β .

The pseudocode of Algorithm 29.2.2 is referred to Fletcher-Reeves definition of β^i $\beta^i = \|\nabla f(\mathbf{x}^i)\| / \|\nabla f(\mathbf{x}^{i-1})\|^2$.

This algorithm converges in at most n iterations.

ALGORITHM 29.2.2 Pseudocode for conjugate gradient method for arbitrary functions

```

1: procedure CGA( $Q, q, x, \varepsilon$ )
2:    $\nabla f^- = 0;$ 
3:   while ( $\|\nabla f(x)\| > \varepsilon$ ) do
4:     if ( $\nabla f^- = 0$ ) then
5:        $d \leftarrow -\nabla f(x);$ 
6:     else
7:        $\beta = \|\nabla f(\mathbf{x}^i)\|^2 / \|\nabla f^-\|^2;$ 
8:        $d \leftarrow -\nabla f(x) + \beta d^-;$ 
9:     end if
10:     $\alpha \leftarrow \text{AWLS}(f(x + \alpha d));$ 
11:     $x \leftarrow x + \alpha d;$ 
12:     $d^- \leftarrow d;$ 
13:     $\nabla f^- \leftarrow \nabla f(x);$ 
14:   end while
15: end procedure

```

We have three different formulas for β_i , which coincide in the quadratic case.

1. Polak-Ribière: $\beta^i = \frac{[\nabla f(\mathbf{x}^i)^T (\nabla f(\mathbf{x}^i) - \nabla f(\mathbf{x}^{i-1}))]}{\|\nabla f(\mathbf{x}^{i-1})\|^2}$
2. Hestenes-Stiefel: $\beta^i = \frac{\nabla f(\mathbf{x}^i)^T (\nabla f(\mathbf{x}^i) - \nabla f(\mathbf{x}^{i-1}))}{(\nabla f(\mathbf{x}^i) - \nabla f(\mathbf{x}^{i-1}))^T d^{i-1}}$
3. Dai-Yuan: $\beta^i = \frac{\|\nabla f(\mathbf{x}^i)\|^2}{(\nabla f(\mathbf{x}^i) - \nabla f(\mathbf{x}^{i-1}))^T d^{i-1}}$

Some of these algorithms require some hypothesis on the function in order for the conjugate method to converge.

1. Fletcher-Reeves requires $m_1 < m_2 < \frac{1}{2}$ for $(A) \cap (W')$ to work;
2. $(A) \cap (W') \not\Rightarrow d^i$ of Polak-Ribière is of descent, unless $\beta_{PR}^i = \max\{\beta^i, 0\}$.

Sometimes it's important to restart from scratches if the algorithm isn't converging, because many bad choices may lead to a bad result.

The idea of taking the gradient and modify it instead of multiplying by a factor, adding the previous direction.

It's possibile to design hybrids between quasi-Newton and conjugate method.

Lecture 30

22nd of November 2018

A. Frangioni

30.1 Deflected gradient methods

The idea behind this family of algorithms, is to determine the next position x^{i+1} using the gradient and summing to it something else that gives us more information.

This kind of algorithms work also in cases in which the gradient isn't continuous.

This methods use the information about the previous iterations without exploiting properties about the second order derivative.

30.1.1 Heavy ball gradient method

The intuition behind this algorithm may be expressed through the following metaphor: an object is moving in the space and it's subject to a force. We can observe that the heavier the object, the stronger should be the force imposed in order to make it describe a certain trajectory.

In this interpretation, we may define the $(i + 1)$ -th iteration as

$$x^{i+1} \leftarrow x^i - \alpha^i \nabla f(x^i) + \beta^i(x^i - x^{i-1}),$$

where β^i is called **momentum**, x^i **heavy** and $\nabla f(x^i)$ **force**.

This isn't a descent algorithm: we are not choosing a direction and doing line search. We have no guarantee that the value of the function after one iteration will be smaller than the previous one.

First thing to do is to choose α^i and β^i properly.

We can prove for some cases that these methods are better than gradient method, although they aren't as good as Newton or quasi Newton. Their strength resides in their simplicity though.

Notice that if the smaller eigenvalue isn't zero (i.e. quadratic case) we have a close formula to choose α and β independently from the iteration:

$$\alpha = \frac{4}{(\sqrt{\lambda^1} + \sqrt{\lambda^n})^2}, \quad \beta = \max \left\{ \left| 1 - \sqrt{\alpha \lambda^n} \right|, \left| 1 - \sqrt{\alpha \lambda^1} \right| \right\}^2$$

We may observe that the step we take is something that goes like $\frac{1}{L}$, where L is the Lipschitz constant, since λ_n is very small. With these choices the rate is the following. We observe that in the gradient the rate is the same, although there aren't the square roots

$$\|x^{i+1} - x_*\| \leq \left(\frac{\sqrt{\lambda^1} - \sqrt{\lambda^n}}{\sqrt{\lambda^1} + \sqrt{\lambda^n}} \right) \|x^i - x_*\|$$

An alternative idea could be choosing β^i and finding α^i using line search. A possible issue is that we don't know if we are moving along a descending direction, but in this method it is perfectly acceptable not to make any movement at a single step (notice that in gradient method if one step has size 0 then we will not move anymore).

β^i is seen as an hyperparameter, hence its value is tuned rerunning the algorithm several times.

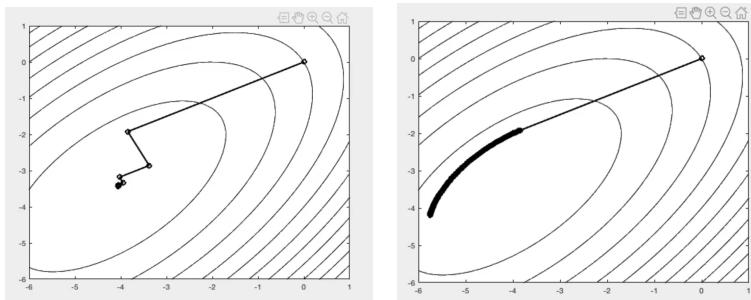


FIGURE 30.1: On the left side Newton method and on the right side the heavy ball method.

The plot of the convergence of the heavy ball method gives a graphical idea of the fact that the direction isn't orthogonal to the one at the previous iteration, since we have the gradient plus some quantity.

In particular, the bigger β^i the "less orthogonal" the steps are. This feature allows the algorithm to have good performances on elongated functions.

30.1.2 Accelerated gradient

This method works only on convex functions, it has some similarities with heavy ball, but it's slightly different.

ALGORITHM 30.1.1 Pseudocode for accelerated gradient method.

```

1: procedure ACCG( $f, \nabla f, x, \varepsilon$ )
2:    $x_- \leftarrow x;$ 
3:    $\gamma \leftarrow 1;$ 
4:   repeat
5:      $\gamma_- \leftarrow \gamma;$ 
6:      $\gamma \leftarrow (\sqrt{4\gamma^2 + \gamma^4} - \gamma^2)/2;$ 
7:      $\beta \leftarrow \gamma(1/\gamma_- - 1);$ 
8:      $y \leftarrow x + \beta(x - x_-);$ 
9:      $g \leftarrow \nabla f(y);$ 
10:     $x_- \leftarrow x;$ 
11:     $x \leftarrow y - (1/L)g;$ 
12:   until ( $\|g\| > \varepsilon$ )
13: end procedure

```

The rational behind this algorithm is the following: When we are in a certain point at a certain iteration, we go on a little bit β^i and we end up in a point y . The gradient is computed in that point and used to choose the next point.

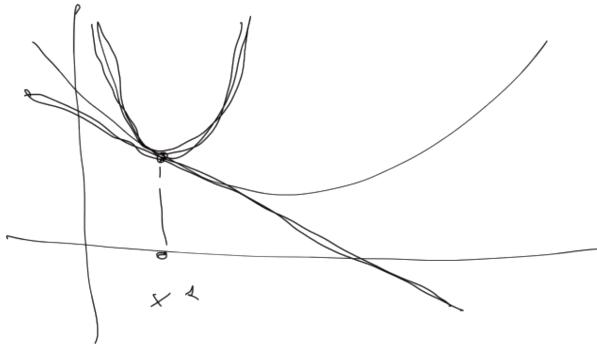


FIGURE 30.2: We build a linear model, which is a lowerbound for the function, then we build a quadratic model, which is above our function.

If we choose γ optimally then the quadratic approximation is very close to the function. We want to find the value for γ s that gives best results in the worst case.

We can prove that the error $\|f(x^i) - f_*\| \leq \sigma^i(f(x^i) - f_*) \searrow 0$ is multiplied by this factor δ_i , that goes like the inverse of i^2 .

Notice that if we choose α small, it will always be small.

The convergence is sublinear.

Theorem 30.1.1 (Optimality of accelerated gradient). *If the function isn't*

strongly convex no algorithm has better convergence than $|f(x^i) - f_| = 3L \frac{\|x^i - x_*\|^2}{32(i+1)^2}$.*

Observation 30.1.1. *This theorem tells us that this algorithm never gets worse than $|f(x^i) - f_*| = 3L \frac{\|x^i - x_*\|^2}{32(i+1)^2}$, but this doesn't imply that this method is fast on average. The state of the art provides a lot of different formulas for β , which of the ones leads to some theoretical results.*

From now on we will move towards a different family of functions, that aren't even differentiable, hence we can't compute the gradient.

30.2 Incremental gradient methods

This method has good performances in real world machine learning cases, where the function is differentiable but we do not want to compute the gradient.

Let $I = \{1, \dots, m\}$ be the set of observations, $X = [X^i \subset \mathcal{X}]_{i \in I}$ the set of inputs and $y = [y^i]_{i \in I}$ the set of outputs. Our goal is to explain y from X .

Since these vectors are uniformly distributed over the space (at least this is our hypothesis) when they get summed we expect some of them to cancel out.

The idea is to rewrite the problem as learning a linear function in the feature space, formally $\min \{ \sum_{i \in I} l(y^i, \langle \Phi(X^i), w \rangle) : w \in \mathbb{R}^n \}$, where $l(\cdot, \cdot)$ is called **loss function**.

We are now interested in computing the gradient, which is not hard to compute since the function is linear: $\nabla f(w) = \sum_{i \in I} \nabla f^i(w) = \sum_{i \in I} -A^i(y^i - A^i w)$.

The issue here is that computing the gradient, although it's the gradient of a very simple function, takes too long to be computed (since there are too many vectors in machine learning datasets).

To overcome this problem we choose to restrict to only a subset of observations. How can we choose such set? Randomly, of course.

At this point the algorithm isn't deterministic anymore, but it's completely **stochastic**.

The intuition behind this algorithm is to take only one observation, compute what is needed on this observation and make a small step.

An online application may be a sensor that produces hundreds of outputs per second and it's not possible to store each of them. They should be used to infer some information and then thrown away.

We study the converge of this kind of algorithms from a stochastic point of view.

In machine learning we always need some regularization, because the tuning of hyperparameters clearly takes into account only the error in the samples that have been seen. Let us regularize the model as follows

$$\min \left\{ \sum_{i \in I} l(y^i, \langle \Phi(X^i), w \rangle) + \mu \Omega(w) : w \in \mathbb{R}^n \right\}$$

The usage of regularization may be useful, since we want to keep close to the minimum, but not reach it. It's enough to change slightly the problem and then solve it.

The regularization hyperparameter $\Omega(w)$ may be chosen as follows:

1. Lasso regularizer (best known): $\Omega(w) = \|w\|_1$;
2. In order to increase sparsity: $\Omega(w) = \|w\|^2$;
3. Leading to feature selection: many $w_j = 0$ as possible.

Ω function is not differentiable, so the function gets non differentiable, as an example look at Figure 30.3, which represents the plot of $f(w_1, w_2) = (3w_1 + 2w_2 - 2)^2 + 10(|w_1| + |w_2|)$.

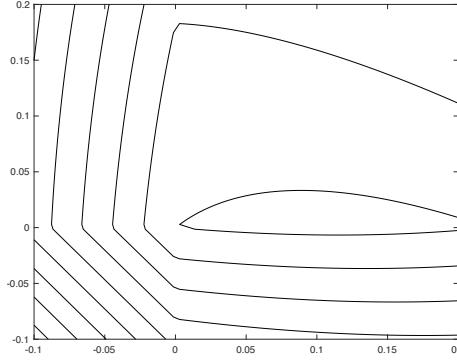


FIGURE 30.3: This function has a lot of kinky points.

30.3 Subgradient methods

These methods are thought for convex functions that are not differentiable.

Let us consider a kinky point: how can we choose between all the subgradients of that point? We assume to be able to compute some subgradients; since the function is convex we may recall

Property 30.3.1. *Let f be a convex function, $\forall g \in \partial f(x), \forall y \in \mathbb{R}^n g(y) \equiv f(y) \geq f(x) + g(y - x)$.*

Let x^* be the optimum, $\langle x^* - x^i, g \rangle$ is smaller than 0. This means that the angle between x and x^* is acute. If we knew the exact direction $x - x^*$ the line search would land on x^* . For a pictorial representation see Figure 30.4

The intuition behind this algorithm is to move in the direction of $-g$, but with a small α^i , because if the step is too large we may end up in a point which is actually further from x^* than the previous step. In that point we may

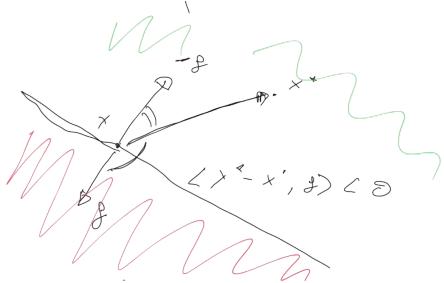


FIGURE 30.4: Since we know that $\langle g, x^* \rangle$ is negative we get that the angle between g and x^* is larger than 90° . Moreover, we know where the optimum is not (red region), hence we restrict to the green region. At this point we will perform a line search on g direction. Since the function is not smooth, the line search may not succeed since the value of the function along the half line from x in g direction may remain the same, but there may be some points there that are closer to x^* .

find a point where perform line search, because that point isn't kinky. In this context we won't try to minimize $\|f(x) - f(x^*)\|$, but we will minimize $\|x - x^*\|$, because the function may zigzag near that point. It goes without saying that choosing a too small value for α leads to a too slow convergence speed.

Lecture 31

28th of November 2018

A. Frangioni

31.1 Subgradient methods

We are still in the hypothesis of convex objective functions that are not differentiable in the whole domain.

💡 Do you recall?

Property 30.3.1 of last lecture: for f convex function, $\forall g \in \partial f(x), \forall y \in \mathbb{R}^n g(y) \equiv f(y) \geq f(x) + g(y - x)$. This is a characterization of the function with respect to the model.

Let us assume we know the minimum point x_* . We observe that the scalar product between the subgradient and the direction we should choose is negative. Formally, $f(x_*) \geq f(x) + \langle g, x_* - x \rangle$, hence $\langle g, x_* - x \rangle \leq f(x_*) - f(x) \leq 0$.

We want to bound the distance between the point at the “next step” and the optimum:

$$\begin{aligned} \|x^{i+1} - x_*\|^2 &\stackrel{(1)}{=} \|x^i - \alpha^i d^i - x_*\|^2 \\ &= \|x^i - x_*\|^2 + 2\alpha^i g^i \frac{(x_* - x^i)}{\|g^i\|} + (\alpha^i)^2 \\ &\stackrel{(2)}{\leq} \|x^i - x_*\|^2 - 2\alpha^i \frac{(f(x^i) - f(x_*))}{\|g^i\|} + (\alpha^i)^2 \end{aligned} \quad (31.1.1)$$

Where $\stackrel{(1)}{=}$ follows from the definition of a normalized step: $x^{i+1} = \frac{x^i - \alpha^i g^i}{\|g^i\|}$ and $\stackrel{(2)}{\leq}$ follows from the inequality we stated above.

Observation 31.1.1. *The distance from the optimum is bounded by a square function in α_i (the step size), where the linear part is negative and the quadratic part is positive.*

Hence, if the steps are short enough we get close to the optimum fast enough, because the linear part dominates the quadratic one.

Formally, $-2 \frac{(f(x^i) - f(x_*))}{\|g^i\|} < 0 + \alpha^i \searrow 0_+$, hence $(\alpha^i)^2 \searrow 0_+$.

An attentive reader may notice that from Equation (31.1.1) follows $\|x^{i+1} - x_\|^2 < \|x^i - x_*\|^2 + (\alpha_i)^2$ and $\|x^{i+1} - x_*\|^2 \leq \|x^1 - x_*\|^2 + \sum_{k=1}^i (\alpha^k)^2$.*

The point here is that we cannot choose a too small α , recall Armijo conditions.

If the series of the squares of step sizes does not diverge, then the sequence does not diverge as well, hence it converges somewhere, say \bar{x} .

The convergence of the series of the squares may be obtained using the following

Definition 31.1.1 (Diminishing-Square Summable). *We term a series that diverges, while the series of the squares of the terms diverges, as **diminishing-square summable**.*

Formally,

$$(DSS) \sum_{i=1}^{\infty} \alpha^i = \infty \wedge \sum_{i=1}^{\infty} (\alpha^i)^2 < \infty.$$

Assumptions: function convex, definite in all its domain, hence the norm of the gradient will never become very large.

Fact 31.1.1. *We claim that the sequence of the stepsizes is a diminishing-square sequence.*

Proof by contraddiction. Let us assume $f(x^i) - f_* \geq \varepsilon > 0$, $\forall i$. Then,

$$\|x^{i+1} - x_*\|^2 \leq \|x^1 - x_*\|^2 - \delta \cdot \sum_{k=1}^i \alpha^k + \sum_{k=1}^i (\alpha^k)^2$$

The contraddiction is due to the fact that as $i \rightarrow \infty$ the right-hand side goes to $-\infty$. \square

This family of algorithms is clearly incredibly robust in theory, but in practice it does not work very well.

Let us make an experiment: let us suppose we know the minimum of the function f .

Definition 31.1.2 (Plyak stepsize). *Let f be our convex objective function and let x_* be the optimum for f . We term **Polyak stepsize**:*

$$(PSS) \alpha^i = \beta^i \frac{(f(x^i) - f(x_*))}{\|g^i\|}$$

where $\beta \in (0, 2)$.

If we pick a Polyak stepsize, then $\|x^{i+1} - x_*\|^2 < \|x^i - x_*\|^2$, so $\{x^i\} \rightarrow x_*$. The best value for β_i is 1. In this case, we obtain what follows by substituting $\beta_i = 1$ in Equation (31.1.1)

$$\frac{(f(x^i) - f(x_*))^2}{\|g^i\|^2} \leq \|x^i - x_*\|^2 - \|x^{i+1} - x_*\|^2$$

The problem is that we don't know the optimum.

Let us assume that we know it and compute the efficiency:

Since we know that the sequence is bounded, we know that the objective function is globally Lipschitz (the norm of the gradient is bounded above).

The point is that the sequence $\{x_1\}$ is not necessarily monotone, so we pick the so-called record value of best value ($\underline{f}^i = \min\{f(x^h) : h = 1, \dots, i\}$)

$$\frac{(\underline{f}^i - f(x_*))^2}{L^2} \leq \frac{(f^i - f(x_*))^2}{\|g_i\|^2} \leq \|x^i - x_*\|^2 - \|x^{i+1} - x_*\|^2$$

Summing for $i = 1, \dots, k$ we obtain a telescopic series:

$$\|x^1 - x_*\| - \|\cancel{x}^2 - x_*\| + \|\cancel{x}^2 - x_*\| - \|\cancel{x}^3 - x_*\| + \dots + \|\cancel{x}^k - x_*\| - \|x^{k+1} - x_*\|$$

Hence resulting in

$$k \cdot \frac{(\underline{f}^k - f(x_*))^2}{L^2} \leq \|x^1 - x_*\|^2 - \|x^{k+1} - x_*\|^2 \leq \|x^1 - x_*\|^2 = R$$

which is equivalent to $(\underline{f}^k - f(x^*))^2 \leq \frac{R^2 L^2}{k}$, which is again equivalent to $\underline{f}^k - f(x^*) \leq \sqrt{\frac{RL}{k}}$, where L is the Lipschitz constant.

The issue here is that the convergence is sublinear: $k \geq \frac{1}{\varepsilon^2}$.

Theorem 31.1.2. *Take an algorithm that uses only the subgradient. It's possible to construct a function that makes the algorithm converge with sublinear speed. Hence, we cannot do better.*

It comes without saying that although this algorithm is not very good it is the "less bad" it can be.

There are some lucky cases in which we do know the optimal value, but this is not the case usually.

31.1.1 Target level stepsize

Let us assume $f(x_*)$ is unknown. The only information available is that this value is below any value in any iteration.

The rationale behind this algorithm is to assume to know the optimal value and as soon as we realize it is not correct we change it.

Let us first give an informal description of the algorithm:

- δ is the displacement: how much below the function is with respect to the best value obtained so far;
- reference value $f_{rec} = f$. At the beginning is the value at the first iterate and then we define the target value as the difference between the reference value and some δ (at the beginning δ_0).

ALGORITHM 31.1.1 Pseudocode for target level stepsize.

```

1: procedure SGPTL( $f, g, x, i_{max}, \beta, \delta_0, R, \rho$ )
2:    $r \leftarrow 0$ ;
3:    $\delta \leftarrow \delta_0$ ;
4:    $f_{ref} \leftarrow f_{rec} \leftarrow f(x)$ ;
5:    $i \leftarrow 1$ ;
6:   while ( $i < i_{max}$ ) do
7:      $g = g(x)$ ;
8:      $\alpha = \beta(f(x) - (f_{ref} - \delta)) / \|g\|^2$ ;
9:      $x \leftarrow x - \alpha g$ ;
10:    if ( $f(x) \leq f_{ref} - \delta/2$ ) then
11:       $f_{ref} \leftarrow f_{rec}$ ;
12:       $r \leftarrow 0$ ;
13:    else
14:      if ( $r > R$ ) then
15:         $\delta \leftarrow \delta\rho$ ;
16:         $r \leftarrow 0$ ;
17:      else
18:         $r \leftarrow r + \alpha \|g\|$ ;
19:      end if
20:    end if
21:   end while
22:    $f_{rec} \leftarrow \min\{f_{rec}, f(x)\}$ ;
23:    $i \leftarrow i + 1$ ;
24: end procedure

```

At this point we defined the algorithm and we are ready to implement it, except for the fact that we need to choose of a lot of parameters. A way to choose them is to use the ML approach: try many possible values.

Two big issues of this algorithm are that it does not provide a good stopping criterion and it is very sensitive to many parameters.

31.2 Deflected subgradient

The idea is to use the same trick of ball-step (also called primal-dual).

Let us assume that our function was differentiable. The subgradient method collapses to the gradient method and we know that the gradient method does not provide a good convergence. Yet, deflection is possible: $d^i = \gamma^i g^i + (1 -$

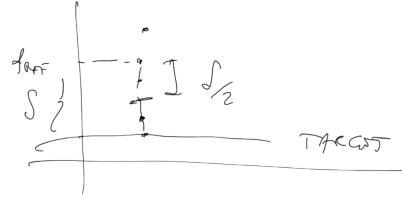


FIGURE 31.1: There are two cases: either the function value is significantly below the reference (for example $\frac{\delta}{2}$ below the reference) or it's not. If we are in the “happy case” (we just move the reference to the best value). For what concerns the “unhappy case”, if we are unhappy for 1 iteration, no problem. Two iterations? no problem. Many iterations? Problem: after some iterations in which we are not improving it means that we have to decrease the reference values. How? r is updated at each bad step and reset when a good step occurs. When r gets too large we decrease δ and reset everything.

$\gamma^i) d^{i-1}$, $x^{i+1} = x^i - \alpha^i d^i$. We can prove that d^i approximates the subgradient. We can also prove that the algorithm converges in the end. The parameters of this algorithm are two: β (stepsize) and γ (deflection). In order to choose them we have two different approaches:

STEP SIZE-RESTRICTED \equiv deflection-first. We first choose β and when choosing α we need to take into account β : $\alpha^i = \frac{\beta^i(f(x^i) - f_*)}{\|d^i\|} \wedge \beta^i \leq \gamma^i$ “as deflection \nearrow , step size has to \searrow ”;

DEFLECTION-RESTRICTED \equiv stepsize-first. We first choose γ , then we pick a step size that depends on γ :

$$(DSS) \wedge \frac{\alpha^{i-1} \|d^{i-1}\|}{(f(x^i) - f_*) + \alpha^{i-1}} \|d^{i-1}\| \leq \gamma^i$$

“as $f(x^i) \rightarrow f_*$, deflection \searrow ”

This algorithm gets the optimal $O(1/\varepsilon^2)$ on average, sadly not worst case.

31.3 Smoothed gradient methods

Let us assume that the target function is a **Lagrangian function**.

Definition 31.3.1 (Lagrangian function). *Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ be a function of the following shape:*

$$f(x) = \max\{x^T A z : z \in Z\}$$

where Z is convex and bounded.

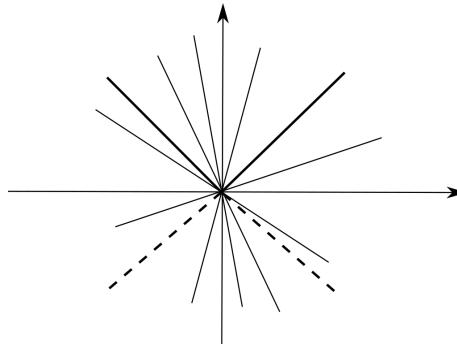


FIGURE 31.2: Let us take the absolute value function $f(x) = |x|$. This function would be differentiable, if it wasn't for some nasty points (in this case 0, were the optimization problem has many optimal solutions). In 0 there are many subgradients, so it has many optima.

Let us assume that Z is also compact.

A graphical example in the case of $f(x) = |x| = \max\{x, -x\} = \max\{zx : z \in [-1, 1]\}$ is shown in Figure 31.2.

In the case of the absolute value, the nasty trick is “to make it have only one optimal solution” in the point 0.

This result is obtained adding a small quadratic term: $f(x) = \max\{x^T Az - \mu \|z\|^2 : z \in G\}$ that is shown in Figure 31.3. At this point the new function f_μ is not the original function anymore, but it is very close to it whenever the μ is small.

Notice that this new function is smooth (differentiable).

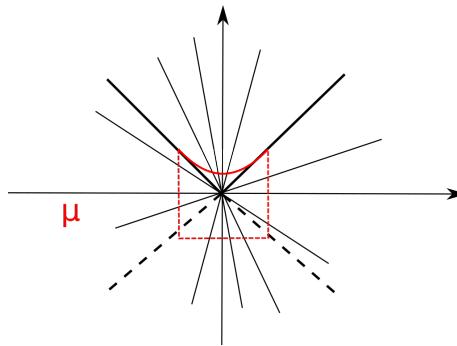


FIGURE 31.3: Geometric intuition of the usage of variable μ .

It might be not very easy to compute once chosen the value for μ .

We are solving a problem which is different from the original one and would be exactly the same if $\mu = 0$.

On the other hand for $\mu = 0$ we have the problem which is not differentiable once again, so we need to keep close to 0 but not too close.

At this point we are in the situation of $f_\mu(x) \leq f(x) \leq f_\mu(x) + \mu R$, such that as $\mu \searrow 0$, “ $\operatorname{argmin} \{f_\mu(x)\} \rightarrow x_*$ ”.

The new function is not only convex, but it is also Lipschitz continuous: ∇f_μ Lipschitz with $L = \|A\|^2/\mu$, but it is “less and less Lipschitz” as $\mu \searrow 0$.

Fact 31.3.1. *If $f_* > -\infty$ and picking a very special value of μ ($\mu = \varepsilon/(2R)$), then an appropriate ACCG obtains $f(x^i) - f_* \leq \varepsilon$ for $i \geq 4 \cdot \|A\| \cdot \|x_*\| \cdot \sqrt{R}/\varepsilon$.*

We observe that the convergence is much better, because it depends on $O(1/\varepsilon)$ instead of $O(1/\varepsilon^2)$.



FIGURE 31.4: At the beginning we will make a lot of bad steps (the upper gray line). We can improve (pick the black line) changing the ε value and we obtain something that looks more stable. The more precision we want, the smaller the step we make. At the ending it pays, but at the beginning it is not so.

31.4 Cutting-plane algorithm

We cheated to get first order information, we want to do more. We want to cheat and have also the second order information.

We want to use the same idea of limited memory quasi Newton methods. Using some limited memory of the Hessian, in order to understand the curvature.

The point is that the directional derivatives are defined and (if computed massively) give me a hint of the curvature of the function (cfr. Figure 31.5). Notice that it is not possible to build a matrix, because it is not defined.

Let us say we have performed i iterates, then we have collected i subgradients ($\mathcal{B} = \{(x^i, f^i = f(x^i), g^i \in \partial f(x^i))\} \equiv$ bundle of first-order information) and function values.

We can now define a piece wise linear function defined as the maximum of the first order model:

$$f_{\mathcal{B}}(x) = \max\{f^i + g^i(x - x^i) : (x^i, f^i, g^i) \in \mathcal{B}\}$$

At this point we can apply Newton method: first we minimize the model and the use the minimum as next point. We collect information in that specific point and then we repeat.

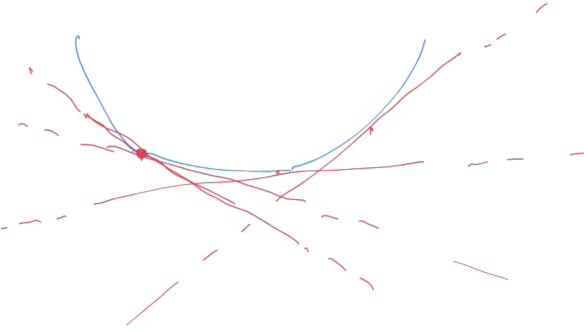


FIGURE 31.5: The idea is that we do not compute a subgradient and discard it. We store it, thus resulting in a model

Notice that the model is always below the objective function ($f_{\mathcal{B}}(x) \leq f(x) \forall x$), hence $\min\{f_{\mathcal{B}}(x)\} \leq f_*$, so $x_{\mathcal{B}}^* \in \operatorname{argmin}\{f_{\mathcal{B}}(x)\} \approx x_*$.

This function has many kinky points, so how to minimize it? Dirty trick from Ricerca Operativa:

$$\min\{f_{\mathcal{B}}(x)\} = \min\{v : v \geq f^i + g^i(x - x^i), \text{ where } (x^i, f^i, g^i) \in \mathcal{B}\}$$

And on this problem, we can use the simplex method.

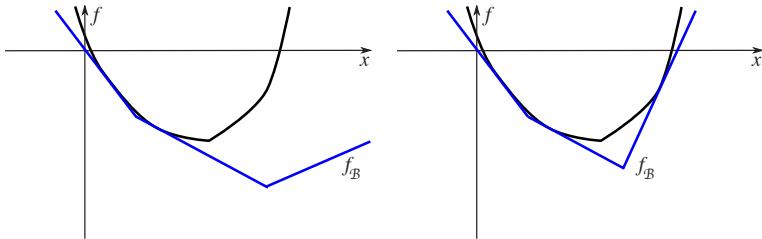


FIGURE 31.6: A geometric representation of how the model (blue line) changes after one iteration.

At this point we have obtained both an upperbound and a lower bound, which go one towards the other.

We may decide to stop iterating when they are “close enough”.

The problem is that at each step we need to solve a minimization problem and the convergence is very slow.

It's also possibile that the blue function (the model) does not have a minimum (unbounded below).

How to overcome this problem? Use so-called bundle methods.

31.5 Bundle methods

The intuition behind this is that we want to keep quite close to the point where the model corresponds to the function, because the further we get the more the difference from the function.

A way to express this is to add a quadratic quantity to the function that grows when we move outside the current point.

Definition 31.5.1 (Stabilized master problem). *We term **stabilized master problem** the following*

$$\min\{f_{\mathcal{B}}(x) + \mu\|x - \bar{x}\|^2/2\}$$

This improved model cannot be undounded below (quadratic function), but we need to choose μ and \bar{x} wisely.

A possible way is to move the **stability center** whenever the current value is better than the best encountered so far.

ALGORITHM 31.5.1 Pseudocode for boundle method.

```

1: procedure PBM( $f, g, \bar{x}, m_1, \varepsilon$ )
2:   choose  $\mu$ ;
3:    $\mathcal{B} \leftarrow \{(\bar{x}, f(\bar{x}), g(\bar{x}))\}$ ;
4:   while ( true ) do
5:      $x^* \leftarrow \operatorname{argmin} \{f_{\mathcal{B}}(x) + \mu\|x - \bar{x}\|^2/2\}$ ;
6:     if ( $\mu\|x^* - \bar{x}\|_2 \leq \varepsilon$ ) then
7:       break;
8:     end if
9:     if ( $f(x^*) \leq f(\bar{x}) + m_1(f_{\mathcal{B}}(x^*) - f(\bar{x}))$ ) then
10:       $\bar{x} \leftarrow x^*$ ;
11:      possibly decrease  $\mu$ ;
12:    else
13:      possibly increase  $\mu$ ;
14:    end if
15:     $\mathcal{B} \leftarrow \mathcal{B} \cup (x^*, f(x^*), g(x^*))$ ;
16:   end while
17: end procedure

```

This algorithm may never move (without cycling, luckily), but at least we gained some information.

The bundle method converges in few steps, although each step is quite costly.

We reached a point where to solve an unconstrained problem we need to solve a constrained one, so from next lecture we will start dealing with constrained optimization problems.

Lecture 32

30th of November 2018

A. Frangioni

32.1 Constrained optimization

In this lecture we address the problem of finding the **optimum** of a function in a subset of its domain, called X . The term optimum differs from the minimum, because the optimum in that subset may not be a minimum of the whole function.

$$f_* = \min\{f(x) : x \in X\}$$

Definition 32.1.1 (Local optimum). *Given a function f and a constraint set X , we denote **local optimum** the point where the function assumes the minimum value inside the set X . Formally, $\min\{f(x) : x \in \mathcal{B}(x_*, \varepsilon) \cap X\}$ for some $\varepsilon > 0$.*

Notice that the only points in which the constraint adds some informations are the ones on the boundary, as shown in Figure 32.1

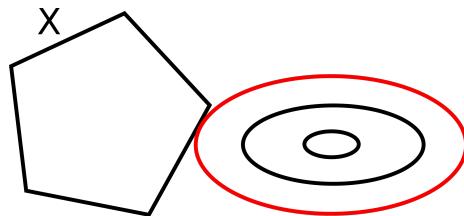


FIGURE 32.1: The red line is level set of the function corresponding to the smallest value that touches the set X . The point in the intersection is not a saddle point of the function f , although it is the minimum.

There are two kinds of constraints:

FAKE ONES: this first kind is such that the minimum of the function lies inside the set X , hence there is no need to use the constraints at all;

REAL ONES: when the optimal is on the boundary. This is the case of linear functions, because the gradient is constant $\nabla f(x) = c$.

At this point we want to decide if a point on the boundary is an optimum. In this context it is important how the boundary is defined.

32.1.1 Linear equality constraints

A constraint of this kind is very simple: it is a subspace, as shown in Figure 32.2.

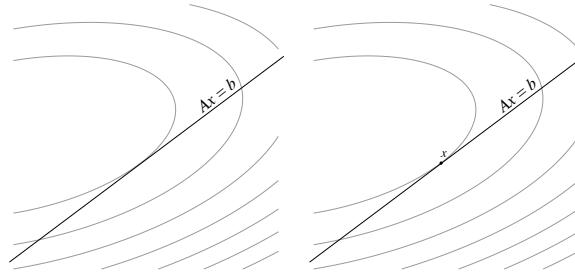


FIGURE 32.2: Linear constraint and a point on the boundary.

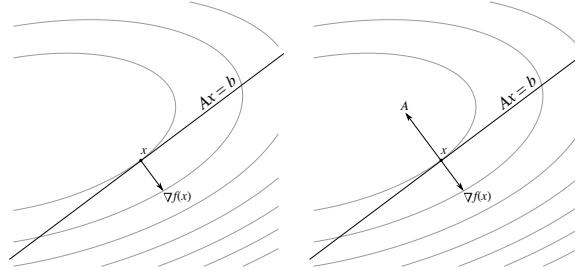


FIGURE 32.3: The gradient is orthogonal to the level set in that point, when the function is smooth. The same holds for matrix A .

$\min\{f(x) : Ax = b\}$, where the rank of A counts the number of linearly independent rows.

Let us assume that there are not linearly independent rows in A , then or this behaviour is reflected in B or the system does not have any solution.

In the case of presence of linearly dependent columns such columns may be eliminated to ease the computation without loss of information.

The intuition behind what follows is that each linear constraint kills one degree of freedom, formally $\det(A_B) \neq 0 \Rightarrow Ax = b \equiv x_B = A_B^{-1}(b - A_N x_N) \Rightarrow$

We want to extract a submatrix A_B from $A \in M(m, n, \mathbb{R})$, such that $A_B \in M(m, \mathbb{R})$ and then the system induces a partitioning in the variables as well.

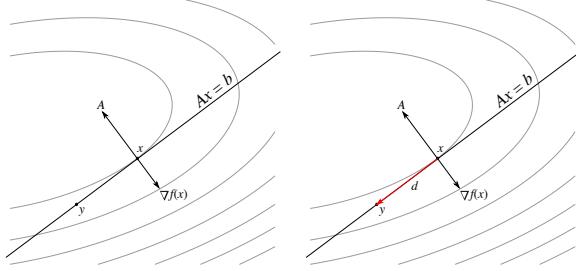


FIGURE 32.4: If we take any other point in the space it has to be orthogonal to A .

$A = [A_B, A_N]$, $x = [x_B, x_N]$, so the system becomes $A_B x_B + A_N x_N = b$ and (since A_B is non singular) $X_B + A_B^{-1} x_N = A_B^{-1} b$ in other words, given the independent variables we can compute the values of the dependent ones and this is a linear operation.

The original optimization problem becomes an optimization problem on a reduced space, formally $\min\{r(w) = f(Dw + d) : w \in \mathbb{R}^{n-m}\}$, where $D = \begin{bmatrix} -A_B^{-1} A_N \\ I \end{bmatrix}$ and $d = \begin{bmatrix} A_B^{-1} b \\ 0 \end{bmatrix}$. For each point in the smaller space we can compute the function in the larger space.

How can we compute the gradient of $r(w)$? The gradient is $\nabla r(w) = D^T \nabla f(Dw + d)$. The fact that w^* is an optimum implies that $\nabla r(w^*) = 0$.

$$D = \begin{bmatrix} -A_B^{-1} A_N \\ I \end{bmatrix} \text{ then } AD = [A_B, A_N] \cdot \begin{bmatrix} -A_B^{-1} A_N \\ I \end{bmatrix} = -A_B \cdot A_B^{-1} \cdot A_N + A_N = 0.$$

Now the point is taking a multiple of matrix A , finding a feasible x and corresponding w (because there is a bijection), then finding the value μ that allows the equality.

Theorem 32.1.1. Let $Ax = b$ be a linear system and w such that $x = [A_B^{-1}(b - A_N w), w]$. If $\exists \mu \in \mathbb{R}^m$ s.t. $\mu A = \nabla f(x)$ then $r(w) = D^T \nabla f(x) = 0$, see Figure 32.5. In other words, it is equivalent to find a stationary point x for the original problem (P) or finding the stationary point w for (R).

Definition 32.1.2 (Poorman's Karush Kuhn-Tucker conditions). A point is a good candidate for being a minimum of the constrained problem if and only if it satisfy **Poorman's KKT conditions**, namely the problem is feasible and that $\exists \mu \in \mathbb{R}^m$ s.t. $\mu A = \nabla f(x)$.

Theorem 32.1.2. Let f be a convex function, then KKT conditions are enough for optimality.

A very naive explanation of the theorem is that if the function is convex also the restriction is convex and a stationary point of a convex function is a minimum.

Our idea is to characterize the directions we can move along in order to find new points that satisfy the constraint. Formally, $Ax = b$ is our constraint

and we want to move towards $x + d$ and stay in the feasible region. How? $A(w + d) = b \Leftrightarrow Ax + Ad = b \Leftrightarrow Ad = 0$, since $Ax = b$. The only way to move along the constraint is choosing a direction which scalar product with A is 0, hence 0 scalar product with the gradient.

From now on we would like to study the behaviour on constrained problems where the constraints are equalities, but inequalities.

In order to do this we need some mathematical background.

32.1.2 Background for linear inequality constraints

Definition 32.1.3 (Tangent cone). We call **tangent cone** of X at x $T_X(x) = t$.

$$\{d \in \mathbb{R}^n : \exists \{z_i \in X\} \rightarrow x \wedge \{t_i \geq 0\} \rightarrow 0 \text{ s.t. } d = \lim_{i \rightarrow \infty} \frac{z_i - x}{t_i}\}$$

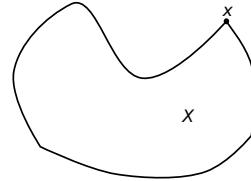


FIGURE 32.5: When the boundary has this shape we can move along directions that point inside the constraints. Tangent directions are not allowed.

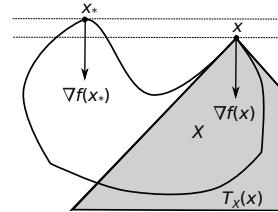


FIGURE 32.6: Geometric representation of tangent cone, which is the region of the space where we can pick the directions. The intuition is to zoom in x and the result of this zooming is a cone.

Theorem 32.1.3. Let \mathcal{C} be a cone, $\forall x \in \mathcal{C} \ \alpha x \in \mathcal{C}, \forall \alpha > 0$.

Theorem 32.1.4. Given a function f , where x is a **local optimum** $\langle \nabla f(x), d \rangle \geq 0 \ \forall d \in T_X(x)$.

Proof. **Proof by contraddiction:** Assume $\exists d \in T_X(x)$ such that $\langle \nabla f(x), d \rangle < 0$, but x is a local optimum.

By definition $\exists X \supset \{z_i\} \rightarrow x$ and $\{t_i\} \rightarrow 0$ such that $d = \lim_{i \rightarrow \infty}^* \frac{z_i - x}{t_i}$.
First order Taylor $f(z_i) - f(x) = \langle \nabla f(x), (z_i - x) \rangle + R(z_i - x)$.

$$\begin{aligned} \lim_{i \rightarrow \infty} \frac{f(z_i) - f(x)}{t_i} &= \lim_{i \rightarrow \infty} \left\langle \nabla f(x), \frac{z_i - x}{t_i} \right\rangle + \frac{R(z_i - x)}{t_i} \\ &= \langle \nabla f(x), d \rangle + \lim_{i \rightarrow \infty} \frac{R(z_i - x)}{t_i} \\ &\stackrel{(1)}{=} \langle \nabla f(x), d \rangle \\ &< 0 \end{aligned} \quad (32.1.1)$$

Where, $\stackrel{(1)}{=}$ follows from $\lim_{i \rightarrow \infty} \frac{R(z_i - x)}{t_i} = 0$ by Taylor. \square

Observation 32.1.1. *The optimum of Theorem 32.1.4 is global when the function is convex, because in that case $X \subseteq x + T_X(x)$. For a geometric idea see Figure 32.7.*

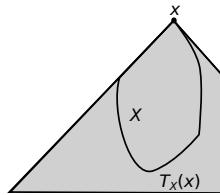


FIGURE 32.7: Convex function.

Observation 32.1.2. *Notice that the rule $\langle \nabla f(x), d \rangle \geq 0 \forall d \in T_X(x) \Rightarrow x$ local optimum does not hold. Let us see a counter example: $\min\{x_2 : x_2 \geq x_1^3\}$, displayed in Figure 32.9.*

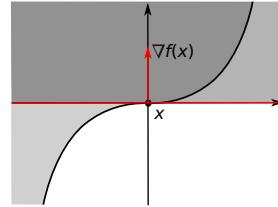


FIGURE 32.8: Let us suppose we pick the direction towards the left part of the function in the saddle point. That direction is promising and actually the value of the function decreases. In this case the problem is that the constraint is not convex.

Now we need a more manageable object for $T_X(x)$.

Definition 32.1.4 (Cone of feasible directions).

*Intuitively, we are in x and we want to find all directions **feasible cone** such that there exist small but not 0 steps such that all the points on this direction are feasible. Formally, a **feasible cone** is $F_X(x) = \{d \in \mathbb{R}^n : \exists \bar{\varepsilon} > 0 \text{ such that } x + \varepsilon d \in X, \forall \varepsilon \in [0, \bar{\varepsilon}]\}$.*

Fact 32.1.5. *The properties of such cone are:*

1. T_X closed, F_X in general not (hence the cone of feasible directions is the tangent cone minus the tangent directions);
2. $\text{cl}(F_X) \subseteq T_X$, where $\text{cl}(F_X)$ is the closure of the cone of feasible directions;
3. If X convex then the cones coincide: T_X and F_X convex and $\text{cl } F_X = T_X$.

We are now interested in finding a better characterization of the cone of feasible directions, hence we introduce a new characterization of the set of constraints X .

FIRST REPRESENTATION:

$$X = \{x \in \mathbb{R}^n : g_i(x) \leq 0 \ i \in \mathcal{I}, h_j(x) = 0 \ j \in \mathcal{J}\}$$

where \mathcal{I} is the set of inequality constraints and \mathcal{J} is the set of equality constraints;

SECOND REPRESENTATION:

$$X = \{x \in \mathbb{R}^n : G(x) \leq 0, H(x) = 0\}$$

where $G = [g_i(x)]_{i \in \mathcal{I}} : \mathbb{R}^n \rightarrow \mathbb{R}^{|\mathcal{I}|}$ and $H = [h_i(x)]_{i \in \mathcal{J}} : \mathbb{R}^n \rightarrow \mathbb{R}^{|\mathcal{J}|}$;

THIRD REPRESENTATION: (hiding equalities)

$$X = \{x \in \mathbb{R}^n : g_i(x) \leq 0 \ i \in \mathcal{I}, h_j(x) \leq 0 \ \wedge \ h_j(x) \geq 0 \ j \in \mathcal{J}\}$$

FOURTH REPRESENTATION: (hiding inequalities into a single function)

$$X = \{x \in \mathbb{R}^n : g(x) = \max\{g_i(x) : i \in \mathcal{I}\} \leq 0 \ i \in \mathcal{I}, h_j(x) = 0 \ j \in \mathcal{J}\}$$

Definition 32.1.5 (Active constraints). *We term **active constraints** at $x \in X$ the following set*

$$\mathcal{A}(x) = \{i \in \mathcal{I} : g_i(x) = 0\} \subseteq \mathcal{I}$$

Let us introduce some useful notation on the subject: let $\mathcal{B} \subseteq \mathcal{I}$ a subset of indices.

We denote $G_{\mathcal{B}} = [g_i(x)]_{i \in \mathcal{B}} : \mathbb{R}^n \rightarrow \mathbb{R}^{|\mathcal{B}|}$ the corresponding set of inequalities.

Definition 32.1.6 (First-order feasible direction cone). We term **First-order feasible direction cone at $x \in X$** :

$$D_X(x) = \{d \in \mathbb{R}^n : \langle \nabla g_i(x), d \rangle \leq 0 \text{ } i \in \mathcal{A}(x) \text{ } \langle \nabla h_j(x), d \rangle = 0 \text{ } j \in \mathcal{J}\} = \{d \in \mathbb{R}^n : (JG_{\mathcal{A}(x)}(x))d \leq 0, (JH(x))d = 0\}$$

Intuitively, the fact that we are looking at the active set means that we are zooming very close to 0.

In this cone we require that all the directions inside it have a negative scalar product with the gradient of the constraints.

A visual example of a first order feasible direction cone is displayed in Figure 32.9.

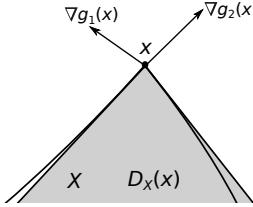


FIGURE 32.9: The first-order feasible direction cone is made of those directions that are orthogonal to the gradient of the constraints $g_1(x)$ and $g_2(x)$ in x .

Fact 32.1.6. The tangent cone is a subset of the first-order feasible direction cone. Formally, $T_X(x) \subseteq D_X(x)$.

We would like the first-order feasible direction cone to be exactly equal to the tangent cone and this is almost always true, except for some pathological cases.

Example 32.1.1. Let our minimum problem be $\min\{\dots : x_1^2 + (x_2 - 1)^2 - 1 \leq 0, x_2 \leq 0\}$.

The plot of such functions is shown in Figure 32.10.

Our claim is that the only feasible point is $X = \{x = [0, 0]\}$.

The only feasible direction is 0, hence cone of feasible direction and the tangent cone are the singleton $\{[0, 0]\}$.

On the other hand, the set of directions that have non-negative scalar product with both g_1 and g_2 are all the x axis.

We would like to ensure we are not in one of these pathological cases and to do this we introduce some conditions.

Fact 32.1.7. The following holds:

AFFINE CONSTRAINTS (AFFC): Let g_i and h_j be affine constraints. Then, $\forall i \in \mathcal{I}$ and $j \in \mathcal{J}$ $T_X(x) = D_X(x) \forall x \in X$.

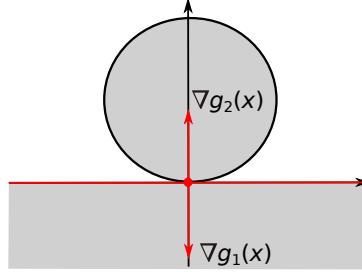


FIGURE 32.10: The circle represents the quadratic constraint, while the semi-plane represents the first degree constraint.

SLATER'S CONDITION (SLAC): Let g_i convex $\forall i \in \mathcal{I}$ and let h_j affine $\forall j \in \mathcal{J}$
 $\exists \bar{x} \in X$ s.t. $g_i(\bar{x}) < 0 \forall i \in \mathcal{I}$. Then $T_X(x) = D_X(x) \forall x \in X$;

LINEAR INDEPENDENCE (LINI): $\bar{x} \in X \wedge$ the vectors $\{\nabla g_i(\bar{x}) : i \in \mathcal{A}(\bar{x})\} \cup \{\nabla h_j(\bar{x}) : j \in \mathcal{J}\}$ linearly independent $\implies T_X(\bar{x}) = D_X(\bar{x})$. Among all these conditions this is the only local one.

It goes without saying that we cannot check all the directions in order to exclude the nasty pathological cases.

Definition 32.1.7 (Dual cone). Let D_X be a **polyhedral cone** $\mathcal{C} = \{d \in \mathbb{R}^n : Ad \leq 0\}$, for some $A \in \mathbb{R}^{k \times n}$.

We term **dual cone** $\mathcal{C}^* = \{c = \sum_{i=1}^k \lambda_i A_i : \lambda \geq 0\}$.

Lemma 32.1.8 (Farka's lemma). Intuitively, this lemma says that pick a vector: either it belongs to the dual cone or there exists a vector in the polyhedral cone which has a negative scalar product with it.

Equivalently, either $c \in \mathcal{C}^*$ or $c \notin \mathcal{C}^*$.

More formally, either $\exists \lambda \geq 0$ s.t. $c = \sum_{i=1}^k \lambda_i A_i$ or $\exists d$ s.t. $Ad \leq 0 \wedge c, d >> 0$.

Theorem 32.1.9 (Karush-Kuhn-Tucker conditions). Let us assume that we found an optimal solution x_* and the constraints qualification holds.

Then $\exists \lambda \in \mathbb{R}_+^{|\mathcal{I}|}$ and $\mu \in \mathbb{R}^{|\mathcal{J}|}$ such that

$$\nabla f(x) + \sum_{i \in \mathcal{A}(x)} \lambda_i \nabla g_i(x) + \sum_{j \in \mathcal{J}} \mu_j \nabla h_j(x) = 0$$

It is interesting to notice that we did not impose $\mu \geq 0$. Let us take an equality constraint $h_j(x) = 0$. This is equivalent to write $h_j(x) \leq 0 \wedge h_j(x) \geq 0$, thus leading to two different multipliers, say λ_j^+ and λ_j^- .

The term of the sum concerning h_j looks like this $\lambda_j^+ \nabla h_j(x_*) - \lambda_j^- \nabla h_j(x_*) = (\lambda_j^+ - \lambda_j^-) \cdot \nabla h_j(x_*)$, where both λ_j^+ and λ_j^- are ≥ 0 , hence their difference (denoted by μ_j) may be either positive or negative.

Fact 32.1.10. *The Karush-Kuhn-Tucker conditions are also written as:*

FEASIBILITY: $x \in X \equiv g_i(x) \leq 0 \quad i \in \mathcal{I}, \quad h_j(x) = 0 \quad j \in \mathcal{J}$

KKT-G: $\nabla f(x) + \sum_{i \in \mathcal{I}} \lambda_i \nabla g_i(x) + \sum_{j \in \mathcal{J}} \mu_j \nabla h_j(x) = 0$

COMPLEMENTARITY SLACKNESS: $\sum_{i \in \mathcal{I}} \lambda_i g_i(x) = 0$

where the second and the third equations formalize the definition of KKT, where the terms of the first sum should be summed only if their constraints are active.

Fact 32.1.11. *Let (P) be a convex problem. In this case, if the Karush-Kuhn-Tucker conditions hold then x is a global optimum.*

Lecture 33

6th of December 2018

A. Frangioni

33.1 Duality

💡 Do you recall?

The KKT-G condition:

$$\nabla f(x) + \sum_{i \in \mathcal{I}} \lambda_i \nabla g_i(x) + \sum_{j \in \mathcal{J}} \mu_j \nabla h_j(x) = 0$$

We would like to understand what is the function of which this is the gradient.

Definition 33.1.1 (Lagrangian function). *Let (P) be our minimum problem over $f : \mathbb{R}^n \rightarrow \mathbb{R}$. We term **Lagrangian function** the following:*

$$L(x; \lambda, \mu) = f(x) + \sum_{i \in \mathcal{I}} \lambda_i g_i(x) + \sum_{j \in \mathcal{J}} \mu_j h_j(x)$$

Fact 33.1.1. *A necessary condition for optimality is that the gradient of the Lagrangian function is 0. Formally, $\nabla L(x; \lambda, \mu) = 0$.*

A first approach to use the Lagrangian function would be taking λ and μ such that the Langrangian is positive semidefinite.

But this is a too strict requirement, in fact we impose:

Definition 33.1.2 (Critical cone). *Let us assume (x, λ, μ) satisfies (KKT). We define the **critical cone** as*

$$C(x, \lambda, \mu) = \left\{ d \in \mathbb{R}^n : \begin{array}{ll} < \nabla g_i(x), d > = 0 & i \in \mathcal{A}(x) \text{ s.t. } \lambda_i^* > 0 \\ < \nabla g_i(x), d > \leq 0 & i \in \mathcal{A}(x) \text{ s.t. } \lambda_i^* = 0 \\ < \nabla h_j(x), d > = 0 & i \in \mathcal{J} \end{array} \right\}$$

Theorem 33.1.2. Let us assume we have a point (x, λ, μ) that satisfies the Karush-Kuhn-Tucker conditions and satisfies the linear independence of the constraints. If x is local optimum then $d^T \nabla_{xx}^2 L(x, \lambda, \mu)d \geq 0 \forall d \in C(x, \lambda, \mu)$.

Informally, if the hypothesis holds, then the Hessian of the Lagrangian function is $\succeq 0$ on the critical cone.

Observation 33.1.1. (x, λ, μ) satisfies $(KKT) \wedge \nabla_{xx}^2 L(x, \lambda, \mu) \succ 0$ on $C(x, \lambda, \mu)$ then x local optimum.

We would like to say something more about λ and μ . Until now we considered the Lagrangian as a function of x , but what if we consider the Lagrangian in terms of λ and μ ?

33.2 Lagrangian duality

Definition 33.2.1 (Lagrangian relaxation). Let us consider the Lagrangian function. We term **Lagrangian relaxation** the function where we fixed λ and μ and minimize on x :

$$\psi(\lambda, \mu) = \min\{L(x, \lambda, \mu) : x \in \mathbb{R}^n\}$$

The relaxation leads to an unconstrained problem and we learnt how to solve one of those.

This Lagrangian relaxation leads to the definition of **lagrangian dual** ψ .

Property 33.2.1. The dual function has the following properties:

- ψ is concave, but $\psi(\lambda, \mu) = -\infty$;
- ψ is non differentiable, although f , g_i and h_j are;
- Let \bar{x} be optimal in $(R_{\lambda, \mu})$, then $[\sum_{i \in \mathcal{I}} \nabla g_i(\bar{x}) + \sum_{j \in \mathcal{J}} \nabla h_j(\bar{x})] \in \partial\psi(\lambda, \mu)$
- \forall fixed $\lambda \geq 0, \mu, \bar{x} \in X \quad \psi(\lambda, \mu) = \min_x L(x, \lambda, \mu) \leq L(\bar{x}, \lambda, \mu) \leq f(\bar{x})$

$$\psi(\lambda, \mu) = \begin{pmatrix} \lambda_1 g_1(\bar{x}) \\ \vdots \\ \lambda_k g_k(\bar{x}) \\ \lambda_1 \mu_1(\bar{x}) \\ \vdots \\ \lambda_p \mu_p(\bar{x}) \end{pmatrix} \text{ is such that } \begin{pmatrix} g_1(\bar{x}) \\ \vdots \\ g_k(\bar{x}) \\ \mu_1(\bar{x}) \\ \vdots \\ \mu_p(\bar{x}) \end{pmatrix} \text{ belongs to the supergradient.}$$

Theorem 33.2.2 (Weak duality). \forall fixed $\lambda \in \mathbb{R}^+ \cup \{0\}, \mu \in \mathbb{R}$ such that $\psi(\lambda, \mu) \leq v(P)$ and let us take any feasible $\bar{x} \in X$ and $g(\bar{x}) \leq 0, h(\bar{x}) = 0$.

It can be proved that $\psi(\lambda, \mu) = \min_x L(x, \lambda, \mu) \leq L(\bar{x}, \lambda, \mu) \leq f(\bar{x})$.

Proof. $L(x, \lambda, \mu) = f(x) + \sum_i \lambda_i g_i(x) + \sum_j \mu_j h_j(x)$. Let us assume we have some $\bar{\lambda}, \bar{\mu}, \bar{x}$, where \bar{x} is feasible (aka $g_i(\bar{x}) \leq 0$ and $h_i(\bar{x}) = 0$ and $\bar{\lambda} \geq 0$).

The value of the dual function is

$$\psi(\bar{\lambda}, \bar{\mu}) = \min_x \{f(x) + \sum_i \lambda_i g_i(x) + \sum_j \mu_j h_j(x)\} \leq f(\bar{x}) + \sum_i \bar{\lambda}_i g_i(\bar{x}) + \sum_j \bar{\mu}_j h_j(\bar{x}),$$

where the last term is cancelled, since $h_i(\bar{x}) = 0$.

On the other hand, $\sum_i \bar{\lambda}_i g_i(x) \leq 0$, which implies that $f(x) \leq 0$. \square

Observation 33.2.1. This theorem gives us the way to prove that a point is an optimum. Let us assume that we found $\lambda_* \leq 0$, x_* and μ_* such that an equality holds: $\psi(\lambda_*, \mu_*) = f(x_*)$. The value of ψ gives us the information about how far we are from the optimum. Let us assume we have $\bar{\lambda}, \bar{\mu}$ and \bar{x} then $\psi(\bar{\lambda}, \bar{\mu}) \leq f(\bar{x}) \leq \psi(\bar{\lambda}, \bar{\mu}) + \varepsilon$ and this tells that we are far from the optimum of a factor ε .

If everything in the original function was convex then the Lagrangian is convex and the computations are easy.

At this point we want to find the biggest lower bound and it may be computed since the function is concave.

$$\max\{\psi(\lambda, \mu) : \lambda \in \mathbb{R}_+^{|\mathcal{I}|}, \mu \in \mathbb{R}^{|\mathcal{T}|}\}$$

Notice that the constraints on λ are very easy, so the problem may be considered somehow unconstrained.

We can use local methods to compute the maximum of this function.

If we are able to compute the gradient of ψ then we have the supergradients of the function f .

Is (P) equal to (D)? Yes, provided that everything is convex. In general the Lagrangian gives a lowerbound.

$$v(D) \leq v(P)$$

Example 33.2.1. Let us take a concave objective function $\min\{-x^2 : 0 \leq x \leq 1\}$, such that its Lagrangian function is $L(x, \lambda) = -x^2 + \lambda_1(x-1) - \lambda_2 x$. It goes without saying that the Lagrangian function is unbounded below (upside-down parabola).

Formally, $\psi(\lambda) = \min_{x \in \mathbb{R}} L(x, \lambda) = -\infty$, $\forall \lambda \in \mathbb{R}^2$, which means that the Lagrangian dual is $v(D) = -\infty < v(P) = -1$.

Theorem 33.2.3. Let us assume that f , g and h are convex, and constraints qualification holds. If the problem has an optimum then the value of the primal and the value of the dual are the same.

$$\text{Formally, } T_X(x_*) D_X(x_*) \Rightarrow v(D) = v(P).$$

Proof. We are assuming x_* is an optimum, so the necessary conditions hold, hence we have the Karush Khun Tucker conditions, then we can find (λ^*, μ^*) that satisfy the KKT with x .

Then our claim is that (λ^*, μ^*) is an optimal solution of the dual (D) and the value of the dual is equal to the value of the primal.

x_* is a stationary point for the Lagrangian function, since it satisfies the KKT conditions then x_* is exactly the minimum, since the Lagrangian function is convex.

Hence, the value of the dual $v(D) \geq \psi(\lambda^*, \mu^*) = L(x_*, \lambda^*, \mu^*) = f(x_*) = v(P) \geq v(D)$. \square

How many solution may $\psi(\lambda, \mu)$ have? In principle many, but what if f is strongly convex, than everything in the Lagrangian is strongly convex, then the solution of ψ is unique and it is also differentiable, but typically not two times differentiable. At this point the Lagrangian dual has a single solution and the optimum of the Lagrangian dual corresponds to an optimum for f .

We only translated a minimum problem on convex functions to another minimum problem on other convex functions. We will see soon that in some cases this approach is advantageous in others it is not.

33.3 Specialized dual

33.3.1 Linead programs

$$(P) \quad \min \{ cx : Ax \geq b \}$$

$$\text{Lagrangian function: } L(x, \lambda) = cx + \lambda(b - Ax) = \lambda b + (c - \lambda A)x$$

Dual function:

$$\psi(\lambda) = \min_{x \in \mathbb{R}^n} L(x, \lambda) = \begin{cases} -\infty & \text{if } c - \lambda A \neq 0 \\ \lambda b & \text{if } c - \lambda A = 0 \end{cases}$$

Since we can find a minimum only on a function of the second case we have that:

$$(D) \quad \max \{ \psi(\lambda) : \lambda \geq 0 \} \equiv \max \{ \lambda b : \lambda A = c, \lambda \geq 0 \}$$

Since the lagrangian is far from having an unique optimal solution the dual does not have a unique maximum so the do not match usually.

33.3.2 Quadratic programs

Notice that the quadratic case is simpler than the linear case.

$$(P) \quad \min \left\{ \frac{1}{2} \|x\|_2^2 : Ax = b \right\} \text{ (linear least-norm solution)}$$

$L(x, \mu) = \frac{1}{2} \|x\|_2^2 + \mu(Ax - b)$, $\nabla_x L = x + \mu A = 0 \iff x = -\mu A$, which is strictly convex since x^2 is strictly convex.

$$\psi(\mu) = \min_{x \in \mathbb{R}^n} L(x, \mu) = L(-\mu A, \mu) = -\frac{1}{2} \mu^T (AA^T) \mu - \mu b$$

Then the Lagrangian dual (D) $\max \left\{ -\frac{1}{2} \mu^T (AA^T) \mu - \mu b : \mu \in \mathbb{R}^m \right\}$

We can generalize with quadratic functions for general problems:

Strictly convex QP: (P)

$$\min \left\{ \frac{1}{2} x^T Q x + q x : Ax \geq b \right\}, Q \succ 0 \implies (D) \quad \max \left\{ \lambda b - \frac{1}{2} v^T Q^{-1} v : \lambda A - v = q, \lambda \geq 0 \right\} \text{ strong duality} \equiv v(P) = v(D) \text{ (almost) always holds}$$

33.3.3 Conic program

We can do duals of things that are not quadratic programs. Sometimes we want to have non linear things to be able to draw different shapes.

Definition 33.3.1 (Conic program). *We term **conic program** $\min\{cx : Ax \geq_K b\}$, where $x \geq_K y \equiv x - y \in K$, where K is a convex cone.*

Example 33.3.1. *It is easy to see that the \geq constraints we saw before are a special case of conic program, where the conic is \mathbb{R}_+^n .*

$$-x_1 - 2x_2 \leq 2$$

$$x_1 + x_2 \geq 1$$

$$2x_1 - x_2 \geq 0$$

Let us write $-x_1 - 2x_2 - 2 = s_0$, $x_1 + x_2 - 1 = s_1$ and $2x_1 - x_2 = s_2$, such that $s_0, s_1, s_2 \geq 0$.

We have a mapping from R^n (where the variables live) to \mathbb{R}^m (where the constraints live) and $Ax - b \in K$.

Rather than writing $s_0, s_1, s_2 \geq 0$ we could write $s_0 \geq \sqrt{s_1^2 + s_2^2}$ and this would still be a conic program.

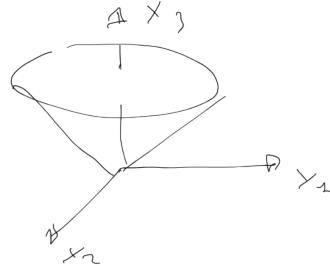


FIGURE 33.1: $x^3 \geq \sqrt{x_1^2 + x_2^2}$ is a convex cone.

The idea is to hide the non linear part in the cone, in particular in \geq_K . Let us see what happens to cones depending on the function.

There are three interesting cones:

- $K = \mathbb{R}_+^n \equiv$ sign constraints \equiv Linear Program;
- $K = \mathbb{L} = \{x \in \mathbb{R}^n : x_n \geq \sqrt{\sum_{i=1}^{n-1} x_i^2}\} \equiv$ Second-Order Cone (or Lorentz cone) Program, that generalizes linear programs;
- $K = \mathbb{S}_+ = \{A \succeq 0\} \equiv$ “ \succeq ” constraints \equiv SemiDefinite Program.

Given the problem that looks like linear except for the cone.

Definition 33.3.2 (Conic dual). *Conic dual: (D) $\max\{yb : yA = c, y \geq_{K^D} 0\}$, where $K^D = \{z : \langle z, x \rangle \geq 0 \forall x \in K\}$ is a dual cone.*

Sometimes constraints qualification does not hold, since the conic program is not linear sometimes.

Explicit form of second order cone program (SOCP) (“explicit data” D_i, d_i, p_i, q_i)
 $\min\{cx : \|D_i x - d_i\|_2 \leq p_i x - q_i, i = 1, \dots, m\}$.

It collapses to a linear program for $D_i = 0$ and $d_i = 0$.

It turns out that the dual has just this explicit form:

$$\max\left\{\sum_{i=1}^m \lambda_i d_i + \nu_i q_i : \sum_{i=1}^m \lambda_i D_i + \nu_i p_i = c, \|\lambda_i\|_2 \leq \nu_i, i = 1, \dots, m\right\}$$

where the ν_i are the dual variables of the rightmost part, while the λ_i are the dual variables of the leftmost part.

We can write the explicit form of semidefinit problems as $\min\{cx : \sum_{i=1}^n x_i A^i \succeq B\}$, where $A^i, B \in M(k, \mathbb{R})$.

There is an even more general form of duality, that we are going to introduce next.

33.4 Fenchel's duality

Definition 33.4.1 (Fenchel's conjugate). We denote **Fenchel's conjugate** of f $f^*(z) = \sup_x \{zx - f(x)\}$.

We can observe that f^* is always convex, even if f is not and it is closed if f is.

The following functions are such that f^* can be computed easily:

1. $f(x) = \frac{1}{2} \|x\|_2^2 \implies f^*(z) = \frac{1}{2} \|z\|_2^2$ (only function s.t. $f^* = f$);
2. $(\|\cdot\|_1)^*(z) = \beta_{B_\infty(0,1)}(z), (\|\cdot\|_\infty)^*(z) = \beta_{B_1(0,1)}(z)$;
3. $f(x) = \max\{g_i x - \alpha_i | i \in I\} \implies f^*(z) = \min \left\{ \sum_{i \in I} \alpha_i \theta_i : \sum_{i \in I} g_i \theta_i = z, \sum_{i \in I} \theta_i = 1, \theta_i \geq 0 | i \in I \right\}$.

Fact 33.4.1. If we are minimizing the sum of two convex functions (P) $\min\{f(x) + g(x)\}$, this problem is the same of maximizing the sum of “almost” the two conjugates: (D) $-\min\{f^*(z) + g^*(-z)\}$.

Lecture 34

12th of December 2018

A. Frangioni

34.1 Quadratic problem with linear equality constraints

Let A be a matrix in $M(m, n, \mathbb{R})$, where $m < n$ (otherwise the system is either fully determined or impossible) and $rk(A) = m$. The constrained quadratic problem may be written as

$$\min \left\{ \frac{1}{2}x^T Qx + qx : Ax = b \right\}$$

where $Q \succeq 0$.

A way to solve this problem is through Karusich Kuhn Tucker system:

$$\begin{array}{ll} (a) & \left[\begin{array}{cc} Q & A^T \\ A & 0 \end{array} \right] \left[\begin{array}{c} x \\ \mu \end{array} \right] = \left[\begin{array}{c} -q \\ b \end{array} \right] \\ (b) & \end{array} \quad (34.1.1)$$

where the first row (a) says that the gradient is a linear combination of the normals of the gradient of the constraints and the second one is just feasibility.

Everything is linear here. This system is symmetric, although indefinite, because it contains many 0s.

We are left with solving the KKT system:

REDUCED KKT: in this method we first add the hypothesis of non-singularity to the matrix Q so we get the following

$$\begin{cases} Qx + A^T \mu = -q & (a) \\ Ax = b & (b) \end{cases}$$

Multiply (a) by AQ^{-1}

$$\begin{cases} A\cancel{Q^{-1}}x + AQ^{-1}A^T\mu = -AQ^{-1}q & \text{(a)} \\ Ax = b & \text{(b)} \end{cases}$$

Multiply (b) by -1 and add to it (a)

$$\begin{cases} Ax + AQ^{-1}A^T\mu = -AQ^{-1}q & \text{(a)} \\ \cancel{Ax} + AQ^{-1}A^T\mu - \cancel{Ax} = -AQ^{-1}q - b & \text{(b)} \end{cases}$$

Multiply (a) by A^{-1}

$$\begin{cases} x + Q^{-1}A^T\mu = -Q^{-1}q & \text{(a)} \\ AQ^{-1}A^T\mu = -AQ^{-1}q - b & \text{(b)} \end{cases}$$

Isolate x

$$\begin{cases} x = -Q^{-1}(A^T\mu + q) & \text{(a)} \\ AQ^{-1}A^T\mu = -AQ^{-1}q - b & \text{(b)} \end{cases}$$

Notice that $0 \preceq AQ^{-1}A^T = M \in M(m, \mathbb{R})$ and may be much smaller than the original one, since its size depends on the number of constraints. The issue here is that the matrix M is less sparse than both A and Q .

NULL SPACE METHOD: In this method we need no assumption on Q .

First of all we rearrange the matrix A in order to have a small square matrix A_B and then the rest of the columns (A_N): $A = [A_B, A_N]$, $x = [x_B, x_N]$, $\det(A_B) \neq 0$.

Replacing A in (b) we get $A_Bx_B + A_Nx_N = b \iff x_B = A_B^{-1} \cdot (b - A_Nx_N)$, hence resulting in $x = Dx_N + d$, where

$$d = \begin{bmatrix} b \\ 0 \end{bmatrix}, D = \begin{bmatrix} -A_B^{-1}A_N \\ I \end{bmatrix} \in M(m, n-m)$$

Notice that D is a basis of the **null space** (or kernel) of A and it is not mandatory to build it like we did above, it is only important to obtain a basis of the kernel.

Let us multiply (a) by D^T and obtain

$$\begin{aligned} D^T(Qx + A^T\mu) &= -D^Tq \\ D^TQx + D^TA^T\mu &= -D^Tq \\ D^TQx + (\cancel{AD})^T\mu &= -D^Tq \\ D^TQ(Dx_N + d) &= -D^Tq \end{aligned} \tag{34.1.2}$$

Where in the last step we applied the definition $x = Dx_N + d$ and hence, $(D^T Q D)x_N = -D^T(Qd + q)$.

We term $H = D^T Q D \in M(n-m, \mathbb{R})$ the reduced Hessian of the problem and notice that whenever the number of constraints is close to the number of variables this matrix is very small.

It is important to note that in order to solve an equality constrained problem we can choose either the reduced KKT method or the null space method, depending on the structure of our problem.

34.2 Inequality constrained problems

The problem, in this case, is written as follows

$$(P) \min\{f(x) : Ax \leq b\}$$

34.2.1 Projected gradient method

The intuition behind this kind of algorithm is that we are interested in finding a direction for the line search not the opposite of the gradient (because then the step size should be put to 0 if we lie on the boundaries, see Figure 34.1), but it is enough to pick a direction which has a negative scalar product with the gradient.

The rationale is to pick the direction that minimizes the norm of the difference with the gradient. Formally

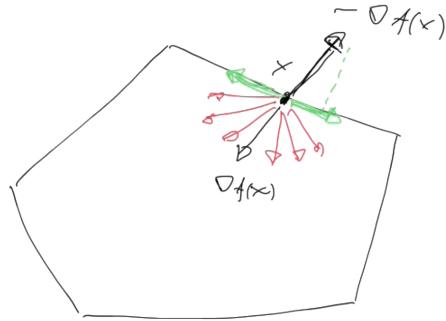


FIGURE 34.1: Geometric idea of how to choose the direction in case of lying on the boundary.

$$d = \operatorname{argmin}\{\|\nabla f(x) - d\|^2 : d \in \mathcal{D}_X(x)\}$$

where $\mathcal{D}_X(x)$ is the feasible cone. Notice that if the difference between the gradient and the direction is 0 it means that we can stop, since the descent direction would bring us outside the feasible set.

A more formal definition of the feasible cone $\mathcal{D}_X(x) = \{d \in \mathbb{R}^n : A_{\mathcal{A}(x)}d \leq 0\}$, where $\mathcal{A}(x)$ is the set of all the active constraints.

For sake of clarity we denote $\bar{A} = A_{\mathcal{A}(x)}$.

At this point we are ready to project the problem in such a way that inequality constraints become equality constraints:

$$\min \left\{ \frac{1}{2} \|\nabla f(x) + d\|^2 = \frac{1}{2} d^T Id + \nabla f(x)d : \bar{A}d = 0 \right\}$$

Note that in this case the Hessian is the identity matrix $I \succ 0$, hence the KKT conditions become simpler:

$$\begin{cases} Qx + A^T \mu = -q \iff d + A^T \mu = -\nabla f(x) \\ Ad = 0 \end{cases}$$

since $x = d$, $A = I$, $b = \nabla f(x)d$, so we get

$$\begin{aligned} d &= -\nabla f(x) - A^T \mu \\ Ad &= -A\nabla f(x) - AA^T \mu \\ AA^T \mu &= -A\nabla f(x) \end{aligned} \tag{34.2.1}$$

resulting in

$$\begin{cases} Qx + A^T \mu = -q \iff d + A^T \mu = -\nabla f(x) \\ Ad = 0 \end{cases}$$

Therefore, we need to solve a system in μ and then we have the direction. If the number of active constraints is small, solving the system is very fast.

Fact 34.2.1. *We can restrict to solve*

$$\begin{cases} \mu = -(\bar{A}\bar{A}^T)^{-1}\bar{A}\nabla f(x) \\ d = (I - \bar{A}^T(\bar{A}\bar{A}^T)^{-1}\bar{A})(-\nabla f(x)) \end{cases}$$

Proof. \bar{A} has full row rank, then is non singular and may be inverted.

$$\begin{cases} \mu = -(\bar{A}\bar{A}^T)^{-1}\bar{A}\nabla f(x) \\ d = -\nabla f(x) + \bar{A}^T(\bar{A}\bar{A}^T)^{-1}\bar{A}\nabla f(x) \end{cases} \iff \begin{cases} \mu = -(\bar{A}\bar{A}^T)^{-1}\bar{A}\nabla f(x) \\ d = (I - \bar{A}^T(\bar{A}\bar{A}^T)^{-1}\bar{A})(-\nabla f(x)) \end{cases}$$

□

If the set of active constraints contains some linear dependent ones, we take the maximal subset of our constraints such that the matrix \bar{A} has full rank.

This procedure is formalized in Algorithm 34.2.1, where at each stage we remove one of the constraints that lead to a negative μ , keeping a set of linearly independent constraints.

In the 14-th line of the algorithm we find that our step size is takes as the minimum step size among the ones that satisfy some subsets of the constraints.

ALGORITHM 34.2.1 Pseudocode for projected gradient method for quadratic functions.

```

1: procedure PGM( $f, A, b, x, \varepsilon$ )
2:   for ( $; ;$ ) do
3:      $B \leftarrow$  maximal  $\subseteq \mathcal{A}(x)$  s.t.  $\text{rank}(A_B) = |B|$ ;
4:     for ( $; ;$ ) do
5:        $d \leftarrow (I - A_B^T (A_B A_B^T)^{-1} A_B)(-\nabla f(x))$ ;
6:       if  $\langle \nabla f(x), d \rangle \leq \varepsilon$  then
7:          $\mu_B \leftarrow -(A_B A_B^T)^{-1} A_B \nabla f(x)$ ;
8:          $\mu_i \leftarrow 0 \forall i \notin B$ ;
9:         if  $\mu_B \geq 0$  then return
10:        end if
11:         $h \leftarrow \min\{i \in B : \mu_i < 0\}$ ;
12:        end if
13:         $\bar{\alpha} \leftarrow \min\{\alpha_i = (b_i - A_i x)/A_i d : A_i d > 0, i \notin B\}$ ;
14:         $x \leftarrow x + \alpha d$ ;
15:        if  $\bar{\alpha} > 0$  then
16:          break;
17:        end if
18:         $k \leftarrow \min\{i \notin B : A_i d > 0 : \alpha_i = 0\}$ ;
19:         $B \leftarrow B \cup \{k\}$ ;
20:      end for
21:       $\alpha \leftarrow \text{Line\_Search}(f, x, d, \bar{\alpha})$ ;
22:       $x \leftarrow x + \alpha d$ ;
23:    end for
24:  end procedure

```

Fact 34.2.2. *The following holds:*

1. $d = 0 \wedge \mu \geq 0 \implies x$ optimal (from KKT);
2. $d = 0 \wedge \exists h \in B$ s.t. $\mu_h < 0 \implies \exists x' \in \{x \in \mathbb{R}^n : A_{B \setminus \{h\}} x = b_{B \setminus \{h\}}, A_h x \leq b_h\}$ s.t. $f(x') < f(x)$;
3. $d \neq 0$ descent direction: $H = I - A_B^T [A_B A_B^T]^{-1} A_B$ symmetric and idempotent $HH = H^T H = H \implies \langle d, \nabla f(x) \rangle < 0$.

Proof.

trivial;

- The intuition is that if we remove h from B next time we will have a descent direction.

$$\exists d \text{ s.t. } A_{B \setminus \{h\}}d = 0 \wedge A_h d < 0 \implies \langle \nabla f(x), d \rangle = \langle -\mu A_B, d \rangle = -\mu_h A_h d < 0;$$

- $\langle d, \nabla f(x) \rangle = \langle -H \nabla f(x), \nabla f(x) \rangle = -\nabla f(x)^T H \nabla f(x) = -(H \nabla f(x))^T H \nabla f(x)$.

□

Once we found the optimal face we move inside that face and we are dealing with a steepest descent, which is slow.

At a certain point of the execution the set B of the active constraints will stabilize and become the one of the optimal solution.

At this point the smart thing to do is to use the KKT conditions, since we know the set of active constraints. This idea is pursued in Section 34.2.3.

34.2.2 Projected gradient method with box constraints

Given the non linear minimum problem of the following shape

$$(P) \min \left\{ f(x) : l \leq x \leq u \right\}$$

a slightly different version of the projected gradient method forces the algorithm to put to 0 the component of the direction which would bring us to go outside the feasible region. This intuition is formalized in Algorithm 34.2.2.

ALGORITHM 34.2.2 Pseudocode for projected gradient method for quadratic functions in the case of box constraints.

```

1: procedure PGMBC( $f, l, u, x, \varepsilon$ )
2:    $d = -\nabla f(x)$ ;
3:    $\bar{\alpha} = \infty$ ;
4:   for ( $i = 1 \dots n$  s.t.  $d_i \neq 0$ ) do
5:     if ( $d_i < 0$ ) then
6:       if ( $x_i = l_i$ ) then
7:          $d_i = 0$ ;
8:       else
9:          $\bar{\alpha} \leftarrow \min\{\bar{\alpha}(x_i - l_i)/d_i\}$ ;
10:      end if
11:    else
12:      if ( $x_i = u_i$ ) then
13:         $d_i = 0$ ;
14:      else
15:         $\bar{\alpha} \leftarrow \min\{\bar{\alpha}(u_i - x_i)/d_i\}$ ;
16:      end if
17:    end if
18:    if ( $\langle \nabla f(x), d \rangle \leq \varepsilon$ ) then return
19:    end if
20:     $\alpha \leftarrow \text{Line\_Search}(f, x, d, \bar{\alpha})$ ;
21:     $x \leftarrow x + \alpha d$ ;
22:  end for
23: end procedure

```

Notice that we can assume that $l_i < u_i \forall i$, because otherwise that component would be fixed.

34.2.3 Active-set method for quadratic programs

Let us be given the following minimum problem

$$\min \left\{ \frac{1}{2} x^T Q x + q x : Ax \leq b \right\}$$

where an important hypothesis is that the problem is quadratic, once we know $\mathcal{A}(x_*)$.

We start from a certain point, such that the constraints are satisfied as equalties. According to KKT conditions, a solution \bar{x} is optimal if \bar{x} is feasible and $\mu \geq 0$. Otherwise, if μ is not positive we eliminate something from the active set and start again. In the case of x unfeasible, we know the descent direction, we only need to revise the step size.

All this machinery is formalized in Algorithm 34.2.3.

ALGORITHM 34.2.3 Pseudocode for active set method for quadratic functions.

```

1: procedure ASMQP( $Q, q, A, b, x, \varepsilon$ )
2:   for ( $B \leftarrow \mathcal{A}(x); ;$ ) do
3:     solve ( $P_B$ )  $\min\{\frac{1}{2}x^T Qx + qx : A_Bx = b_B\}$  for
   ( $\bar{x}, \bar{\mu}_B$ );
4:     if ( $A_i\bar{x} \leq b_i \forall i \notin B$ ) then
5:       if ( $\mu_B \geq 0$ ) then return
6:       end if
7:        $h \leftarrow \min\{i \in B : \mu_i < 0\};$ 
8:        $B \leftarrow B \setminus \{h\};$ 
9:       continue;
10:      end if
11:       $d \leftarrow \bar{x} - x;$ 
12:       $\bar{\alpha} \leftarrow \min\{\alpha_i = (b_i - A_i x) / A_i d : A_i d > 0, i \notin B\};$ 
13:       $x \leftarrow x + \bar{\alpha}d;$ 
14:       $B \leftarrow \mathcal{A}(x);$ 
15:   end for
16: end procedure

```

Notice that this algorithm allows easy solving of the same problem, under box constraints, because we end up having two sets: L , indexes of variables fixed to the lower bound, and U , indexes of variables fixed to the upper bound.

$\{1, \dots, n\} \setminus (L \cup U)$ is the set of the indexes of the free variables.

Thanks to this consideration the problem to solve becomes

$$\min \left\{ \frac{1}{2}x_F^T Q_{FF}x_F + (q_F + u_U^T Q_{UF})x_F \right\} \left[+ \frac{1}{2}x_U^T Q_{UU}x_U + q_U u_U \right]$$

34.2.4 Frank-Wolfe method

Let us take a non-linear function and the following problem

Supposing f is convex we can make a first order model and minimize it over our constraints.

In this case we use the right constraints and an approximation of the function.

ALGORITHM 34.2.4 Pseudocode for Frank-Wolf method for non linear functions.

```
1: procedure FWM( $f, A, b, x, \varepsilon$ )
2:   while ( $\|\nabla f(x)\| > \varepsilon$ ) do
3:      $\bar{x} \leftarrow \operatorname{argmin} \{\langle \nabla f(x), y \rangle : Ay \leq b\};$ 
4:      $d \leftarrow \bar{x} - x;$ 
5:      $\alpha \leftarrow \text{Line\_Search}(f, x, d, 1);$ 
6:      $x \leftarrow x + \alpha d;$ 
7:   end while
8: end procedure
```

Provided that the computation at line 3 of Algorithm 34.2.4 is performed efficiently, then either $d = 0$ and we are in the optimum or $d > 0$ and we are moving toward the optimum.

The linear model is below the function in the quadratic case, hence we get a lower bound.

Lecture 35

14th of December 2018

A. Frangioni

In the previous lecture we addressed the problem of linear constrained optimization. Our first approach was to deal very little with constraints (projected gradient method), after a few improvements we took all of them and modify the function (Frank-Wolfe method).

35.1 Dual methods for linear constrained optimization

In this class of methods constraints are first class citizens.

Let us be given the following optimization problem:

$$\min \left\{ \frac{1}{2}x^T Qx + qx : Ax \leq b \right\}$$

We would like to work with dual feasibility.

\forall fixed $\lambda \geq 0$, this is the lagrangian problem $\psi(\lambda) = \min_x \left\{ \frac{1}{2}x^T Qx + qx + \lambda(b - Ax) \right\} \leq v$ and the Lagrangian dual is the following

$$(D) \max \{\psi(\lambda) : \lambda \geq 0\}$$

which is equivalent to the primal problem.

The solution of the Lagrangian problem is **unique** and this is due to the fact that ψ is concave and $Q \succ 0$. The optimal solution is then $x(\lambda) = Q^{-1}(\lambda A - q)$.

The function is differentiable in the solution, because we have only one sub (or super) gradient which is $\nabla \psi(\lambda) = b - Ax(\lambda)$.

At this point we only need to solve the dual problem, which has the shape of a box constrained optimization, where the box has only one boundary and this can be solved via quasi-Newton methods.

Notice that $\psi \notin C^2$ so the Hessian is not defined.

This dual approach is advantageous in the case of a small number of constraints, because the size of the problem decreases. For example, in the case of one constraint, the Lagrangian dual becomes a problem in one variable, hence solvable through a line search.

In this method the degenerate case (more than one constraint active at a time) is not an issue.

If the quadratic function is convex we may use quasi Newton method. Otherwise the global optimality is not guaranteed and may be used if we accept not to be able to solve the original problem, but only to find a lower bound.

35.1.1 Separable problems and partial dual

Let us assume that our constraints are separable, which means that it is not mandatory to work with all of them, but they can be splitted into constraints of groups of variables.

$$\min\{f(x) : Ax \leq b, Ex \leq d\}$$

We can decide to use a **partial dual**, writing the Lagrangian problem picking only some constraints that we chose:

$$\psi(\lambda) = \min_x \{f(x) + \lambda(b - Ax) : Ex \leq d\}$$

The Lagrangian dual method may be better than projected gradient or worse and it depends on the instance.

In the dual approaches we can't move inside the feasible solution. We find an optimum for the dual, which surely breaks feasibility. Then, if the variable is above the upperbound it gets decreased to the upper bound, otherwise if it is below the lower bound it takes the value of the lower bound.

This way we get an upperbound for the function to be minimized.

35.2 Primal/dual methods or barrier methods

This kind of methods are designed to overcome the cons of dual approaches, namely the fact that ψ does not have the Hessian (and this creates problems to quasi Newton method) and the fact that x is not feasible until the end.

At the same time this methods keep the unconstrained property of the Lagrangian dual.

35.2.1 Barrier function and central path

The rationale behind this algorithm is to minimize a function which penalizes the value of the original function when the solution is getting closer and closer to the boundaries of the feasible set.

$$(P_\mu) \min\{f_\mu(x) = f(x) - \mu \sum_{i=1}^m \log(b_i - A_i x)\}$$

The parameter μ is there to weight the proximity to the boundary.

Property 35.2.1. • If f is convex, f_μ is strictly convex;

- If $f \in \mathcal{C}^2$ then $f_\mu \in \mathcal{C}^2$, since $\log \in \mathcal{C}^\infty$;
- $\forall \mu \exists! x_\mu$ optimal of (P_μ) , since $\mu \sum_{i=1}^m \log(b_i - A_i x)$ is strictly convex;
- As $\mu \rightarrow 0$ x_μ converges to the analytic center of the optimal face. An example of this behaviour may be seen in Figure 35.1.

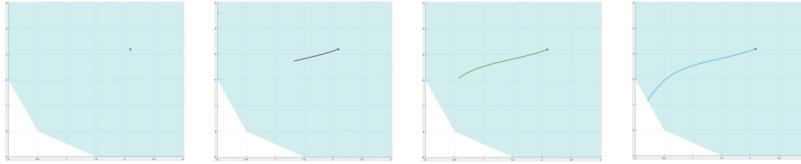


FIGURE 35.1: The trajectory converges to the optimal solution of the problem, when $\mu \rightarrow 0$.

Another interesting property is that, since the barrier function is **self concordant** x^i gets “close” to $x(\mu^i)$ in very few Newton’s steps.

In one step x^{i+1} is “much closer” to $x(\mu^i)$ than x^i was and x^{i+1} is “close” to $x(\mu^{i+1})$, with $\mu^{i+1} \ll \mu^i$ (more formally, $\mu^{i+1} = \tau \mu^i$, $\tau < 1$).

This behaviour may be observed in Figure 35.2

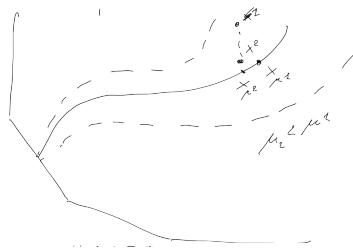


FIGURE 35.2: The dotted line represents a region where the Newton method is very efficient. We are starting from a point x_1 , which belongs to that region and we want to move towards x_{μ^1} . At next iterate x_2 is closer to x_{μ^2} than the current iterate.

The convergence is exponential if τ is very small, but these iterations are very costly, because the Hessian changes at each step, hence it needs to be recomputed.

Let us focus on computing the **Newton's step**.

First, we write the Karush-Kuhn-Tucker conditions:

PRIMAL FEASIBILITY: $Ax + s = b, s \geq 0;$

DUAL FEASIBILITY: $Qx + \lambda A = -q, \lambda \geq 0;$

COMPLEMENTARY SLACKNESS: $\lambda_i s_i = 0, i = 1, \dots, m.$

We can write a slackened version of KKT, imposing $\lambda_i s_i = \mu, i = 1, \dots, m,$ where $\mu \in \mathbb{R}^n$ and should be decreased over iterations until it gets closer enough to 0.

Let us construct $\Lambda, S \in \text{Diag}(m, \mathbb{R})$ such that the diagonal is made of λ_i and s_i respectively.

At this point, we rewrite the problem in terms of the displacement from the fixed current point we are in:

- $x \rightarrow x + \Delta x$
- $s \rightarrow s + \Delta s$
- $\lambda \rightarrow \lambda + \Delta \lambda$

The KKT system becomes:

PRIMAL FEASIBILITY: $Ax + A\Delta x + s + \Delta s = b, s \geq 0;$

DUAL FEASIBILITY: $Qx + Q\Delta x + \lambda A + \Delta \lambda A = -q, \lambda \geq 0;$

COMPLEMENTARY SLACKNESS: $\lambda_i s_i + \lambda_i \Delta s_i + s_i \Delta \lambda_i + \Delta \lambda_i \Delta s_i = \mu, i = 1, \dots, m.$

In this new system of coordinates the first two KKT remain linear, while the third one is no longer linear ($\Delta \lambda_i \Delta s_i$).

$$\begin{bmatrix} Q & A^T & 0 \\ A & 0 & I \\ 0 & S & \Lambda \end{bmatrix} \begin{bmatrix} \Delta x \\ \Delta \lambda \\ \Delta s \end{bmatrix} \stackrel{(1)}{=} \begin{bmatrix} -(Qx + q) - \lambda A \\ b - Ax - s \\ \mu u - \Lambda S u - \Delta \lambda \Delta S u \end{bmatrix} \approx \begin{bmatrix} 0 \\ 0 \\ \mu u - \Lambda S u \end{bmatrix} \quad (35.2.1)$$

Where (1) holds since $\Delta \lambda A = A^T \Delta \lambda^T$, although $\Delta \lambda$ is written without the “transpose” syntax to ease notation.

Notice that $u = \begin{bmatrix} 1 \\ \vdots \\ 1 \end{bmatrix} \in \mathbb{R}^m$ and has the purpose of adjusting dimension:

$$S \Delta \lambda = \begin{bmatrix} s_1 \Delta \lambda_1 \\ \vdots \\ s_m \Delta \lambda_m \end{bmatrix} \in \mathbb{R}^m; \Lambda \Delta s = \begin{bmatrix} \lambda_1 \Delta s_1 \\ \vdots \\ \lambda_m \Delta s_m \end{bmatrix} \in \mathbb{R}^m; \mu u = \begin{bmatrix} \mu \\ \vdots \\ \mu \end{bmatrix} \in \mathbb{R}^m;$$

$$\Lambda S u = \begin{bmatrix} -s_1 \lambda_1 \\ \vdots \\ -s_m \lambda_m \end{bmatrix} \in \mathbb{R}^m; \quad \Delta \Lambda \Delta s u = \begin{bmatrix} -\Delta \lambda_1 \Delta s_1 \\ \vdots \\ -\Delta \lambda_m \Delta s_m \end{bmatrix} \in \mathbb{R}^m;$$

The Newton method can be applied if we discard the non-linear part (written in red), pretending it does not exist. Notice that vector u is the vector of all 1s.

35.3 Primal-dual interior point method

This method is based on the observation that we can solve the dual problem

$$(D) \quad \max \left\{ -\lambda b - \frac{1}{2} x^T Q x : Qx + \lambda A = -q, \lambda \geq 0 \right\}$$

thus obtaining both a lower and upper bound for the solution x .

We term **complementarity gap** $(\frac{1}{2} x^T Q x + qx) - (-\lambda b - \frac{1}{2} x^T Q x) = \lambda s = \mu$.

Once we found a solution for Equation (35.2.1), we perform a step and compute a new couple of primal and dual solutions and reduce the gap μ .

We are left with solving Equation (35.2.1). The trick is to express one between Δs and $\Delta \lambda$ as a linear combination of the other.

For example, let us take the third line of Equation (35.2.1):

$$\begin{aligned} 0\Delta x + S\Delta \lambda + \Lambda \Delta s &= \mu u - \Lambda S u \\ \Lambda \Delta s &= \mu u - \Lambda S u - S\Delta \lambda \\ \Delta s &= \Lambda^{-1} \mu u - \cancel{\Lambda^{-1} \Lambda S u} - \Lambda^{-1} S\Delta \lambda \\ \Delta s &= \Lambda^{-1} \mu u - S u - \Lambda^{-1} S\Delta \lambda \\ \Delta s &= \Lambda^{-1} \cdot (\mu u - S\Delta \lambda) - S u \\ \Delta s &= \Lambda^{-1} \cdot (\mu u - S\Delta \lambda) - s \end{aligned} \tag{35.3.1}$$

We obtain the modified normal equations (or KKT system)

$$\begin{bmatrix} Q & A^T \\ A & -\Lambda^{-1} S \end{bmatrix} \begin{bmatrix} \Delta x \\ \Delta \lambda \end{bmatrix} = \begin{bmatrix} 0 \\ s - \mu \Lambda^{-1} u \end{bmatrix} \tag{35.3.2}$$

where the last row is derived working on the second row of Equation (35.2.1) ($A\Delta x + 0\Delta \lambda + I\Delta s = 0 \Leftrightarrow A\Delta x = -\Delta s$), substituting Equation (35.3.1):

$$\begin{aligned} A\Delta x - [\Lambda^{-1} S]\Delta \lambda &= -\Delta s - [\Lambda^{-1} S]\Delta \lambda \\ &= -\Lambda^{-1}(\mu u - S\Delta \lambda) + s - \Lambda^{-1} S\Delta \lambda \\ &= -\Lambda^{-1} \mu u + \cancel{\Lambda^{-1} S\Delta \lambda} + s - \cancel{\Lambda^{-1} S\Delta \lambda} \\ &= s - \mu \Lambda^{-1} u \end{aligned} \tag{35.3.3}$$

With respect to normal equations of ??, we have in position (2, 2) a quantity $(-\Lambda^{-1}S)$, which is not 0, but it is the opposite of a strictly positive definite matrix.

A possible approach for solving this system is making some calculations and obtain something of the shape of reduced KKT (see Section 34.1).

$$\begin{cases} Q\Delta x + A^T\Delta\lambda = 0 \\ (Q + A^T\Lambda S^{-1}A)\Delta x = A^T(\lambda - \mu S^{-1}u) \end{cases} \quad (35.3.4)$$

where the first set of equations follows from the expansion of the first row of the KKT system (Equation (35.3.2)) and the second one is obtained taking the same row of the same system and substituting the value of $\Delta\lambda$ as follows:

$$\begin{aligned} A\Delta x - \Lambda^{-1}S\Delta\lambda &= s - \mu\Lambda^{-1}u \\ \Lambda^{-1}S\Delta\lambda &= A\Delta x - s + \mu\Lambda^{-1}u \\ \Delta\lambda &= (\Lambda^{-1}S)^{-1}A\Delta x - (\Lambda^{-1}S)^{-1}s + (\Lambda^{-1}S)^{-1}\mu\Lambda^{-1}u \\ \Delta\lambda &= S^{-1}\Lambda A\Delta x - S^{-1}\Lambda s + \mu S^{-1}\Lambda\Lambda^{-1}u \\ \Delta\lambda &= S^{-1}\Lambda A\Delta x - S^{-1}\Lambda s + \mu S^{-1}u \\ \Delta\lambda &= \mu S^{-1}u + \Lambda S^{-1}A\Delta x - \Lambda S^{-1}s \\ \Delta\lambda &= \mu S^{-1}u + \Lambda S^{-1}A\Delta x - \Lambda u \\ \Delta\lambda &= \mu S^{-1}u + \Lambda S^{-1}A\Delta x - \lambda \end{aligned} \quad (35.3.5)$$

Hence,

$$\begin{aligned} Q\Delta x + A^T\Delta\lambda &= 0 \\ Q\Delta x + A^T \cdot (\mu S^{-1}u + \Lambda S^{-1}A\Delta x - \lambda) &= 0 \\ Q\Delta x + A^T\mu S^{-1}u + A^T\Lambda S^{-1}A\Delta x - A^T\lambda &= 0 \\ Q\Delta x + A^T\Lambda S^{-1}A\Delta x &= -A^T\mu S^{-1}u + A^T\lambda \\ (Q + A^T\Lambda S^{-1}A)\Delta x &= A^T(\lambda - \mu S^{-1}u) \end{aligned} \quad (35.3.6)$$

We term $M = Q + A^T\Lambda S^{-1}A$ and the following holds.

Fact 35.3.1. *If A has full column rank (aka it is invertible), $M \succ 0$.*

At this point we need to factorize the matrix M , that changes at each iteration (since ΛS^{-1} does) and this is the bottleneck.

Cholesky factorization may be used, although its complexity is cube. Another downside of this approach is that the matrix M is much denser than A , Λ and S^{-1} .

An orthogonal approach to the reduced KKT is called **predictor-corrector** and it works computing a solution without taking into account the non linear term $\Delta\Lambda\Delta Su$, then computing it according to the approximated solution and repeat until convergence.

The bottleneck again is solving the system in Equation (35.2.1).

For what concerns implementation, we should start from a triplet (x, λ, s) , that could be not feasible and then compute the residuals and iterate until feasibility is reached.

$$r^D = -(Qx + q) - \lambda A, r^P = b - Ax - s.$$

When dealing with the step size we need to highlight the fact that $\lambda + \Delta\lambda \geq 0$, $s + \Delta s \geq 0$ should hold.

In order to achieve this we find the maximum α that satisfies the equality and then multiply it by a constant $\bar{\alpha} = 0.995$ (or 0.9995), in order to get closer.

Let us assume that we also have a bunch of box constraints, hence our problem becomes

$$(P) \quad \min \left\{ \frac{1}{2} x^T Q x + q x : Ax = b, 0 \leq x \leq u \right\}$$

In this special case, things simplify a lot.