J. Frédéric Bonnans, J. Charles Gilbert Claude Lemaréchal, Claudia A. Sagastizábal

# Numerical Optimization Theoretical and Practical Aspects 

May 31, 2006

Springer<br>Berlin Heidelberg New York<br>Barcelona Hong Kong<br>London Milan Paris<br>Tokyo

## Preface

This book is entirely devoted to numerical algorithms for optimization, their theoretical foundations and convergence properties, as well as their implementation, their use, and other practical aspects. The aim is to familiarize the reader with these numerical algorithms: understanding their behaviour in practice, properly using existing software libraries, adequately designing and implementing "home-made" methods, correctly diagnosing the causes of possible difficulties. Expected readers are engineers, Master or Ph.D. students, confirmed researchers, in applied mathematics or from various other disciplines where optimization is a need.

Our aim is therefore not to give most accurate results in optimization, nor to detail the latest refinements of such and such method. First of all, little is said concerning optimization theory itself (optimality conditions, constraint qualification, stability theory). As for algorithms, we limit ourselves most of the time to stable and well-established material. Throughout we keep as a leading thread the actual practical value of optimization methods, in terms of their efficiency to solve real-world problems. Nevertheless, serious attention is paid to the theoretical properties of optimization methods: this book is mainly based upon theorems. Besides, some new and promising results or approaches could not be completely discarded; they are also presented, generally in the form of special sections, mainly aimed at orienting the reader to the relevant bibliography.

An introductory chapter gives some generalities on optimization and iterative algorithms. It contains in particular motivating examples, ranking from meteorological forecast to power production management; they illustrate the large field of branches where optimization finds its applications. Then come four parts, rather independent of each other. The first one is devoted to algorithms for unconstrained optimization which, in addition to their direct usefulness, are a basis for more complex problems. The second part concerns rather special methods, applicable when the usual differentiability assumptions are not satisfied. Such methods appear in the decomposition of large-scale problems and the relaxation of combinatorial problems. Nonlinearly constrained optimization forms the third part, substantially more technical, as the subject is still in evolution. Finally, the fourth part gives a deep account of the more recent interior point methods, originally designed
for the simpler problems of linear and quadratic programming, and whose application to more general situations is the subject of active research.

This book is a translated and improved version of the monograph [43], written in French. The French monograph was used as the textbook of an intensive two week course given several times by the authors, both in France and abroad. Each topic was presented from a theoretical point of view in morning lectures. The afternoons were devoted to implementation issues and related computational work. The conception of such a course is due to J.-B. Hiriart-Urruty, to whom the authors are deeply indebted.

Finally, three of the authors express their warm gratitude to Claude Lemaréchal for having given the impetus to this new work by providing a first English version.

Notes on this revised edition. Besides minor corrections, the present version contains substantial changes with respect to the first edition. First of all, (simplified but) nontrivial application problems have been inserted. They involve the typical operations to be performed when one is faced with a real-life application: modelling, choice of methodology and some theoretical work to motivate it, computer implementation. Such computational exercises help getting a better understanding of optimization methods beyond their theoretical description, by addressing important features to be taken into account when passing to implementation of any numerical algorithm.

In addition, the theoretical background in Part I now includes a discussion on global convergence, and a section on the classical pivotal approach to quadratic programming. Part II has been completely reorganized and expanded. The introductory chapter, on basic subdifferential calculus and duality theory, has two examples of nonsmooth functions that appear often in practice and serve as motivation (pointwise maximum and dual functions). A new section on convergence results for bundle methods has been added. The chapter on applications of nonsmooth optimization, previously focusing on decomposition of complex problems via Lagrangian duality, describes also extensions of bundle methods for handling varying dimensions, for solving constrained problems, and for solving generalized equations. Also, a brief commented review of existing software for nonlinear optimization has been added in Part III.

Finally, the reader will find additional information at http://www-rocq. inria.fr/~gilbert/bgls. The page gathers the data for running the test problems, various optimization codes, including an SQP solver (in Matlab), and pieces of software that solve the computational exercises.

Paris, Grenoble, Rio de Janeiro, May 2006
J. Frédéric Bonnans J. Charles Gilbert

Claude Lemaréchal
Claudia A. Sagastizábal

## Table of Contents

## Preliminaries

1 General Introduction . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 3
1.1 Generalities on Optimization . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 3
1.1.1 The Problem . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 3
1.1.2 Classification . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 4
1.2 Motivation and Examples . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 5
1.2.1 Molecular Biology . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 5
1.2.2 Meteorology . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 6
1.2.3 Trajectory of a Deepwater Vehicle . . . . . . . . . . . . . . . . . 8
1.2.4 Optimization of Power Management . . . . . . . . . . . . . . . . . . 9
1.3 General Principles of Resolution . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 10
1.4 Convergence: Global Aspects . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 12
1.5 Convergence: Local Aspects . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 14
1.6 Computing the Gradient . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 16

Bibliographical Comments . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 19

## Part I Unconstrained Problems

2 Basic Methods . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 25
2.1 Existence Questions . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 25
2.2 Optimality Conditions . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 26
2.3 First-Order Methods . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 27
2.3.1 Gauss-Seidel . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 27
$\begin{array}{ll}\text { 2.3.2 } & \text { Method of Successive Approximations, or Gradient } \\ & \text { Method . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . } \\ 28\end{array}$
2.4 Link with the General Descent Scheme . . . . . . . . . . . . . . . . . . . . 28
2.4.1 Choosing the $\ell_{1}$-Norm . . . . . . . . . . . . . . . . . . . . . . . . . . . . 29
2.4.2 Choosing the $\ell_{2}$-Norm . . . . . . . . . . . . . . . . . . . . . . . . . . . 30
2.5 Steepest-Descent Method . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 30
2.6 Implementation . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 34

Bibliographical Comments . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . . 35
3 Line-Searches ..... 37
3.1 General Scheme ..... 37
3.2 Computing the New $t$ ..... 40
3.3 Optimal Stepsize (for the record only) ..... 42
3.4 Modern Line-Search: Wolfe's Rule ..... 43
3.5 Other Line-Searches: Goldstein and Price, Armijo ..... 47
3.5.1 Goldstein and Price ..... 47
3.5.2 Armijo ..... 47
3.5.3 Remark on the Choice of Constants ..... 48
3.6 Implementation Considerations ..... 49
Bibliographical Comments ..... 50
4 Newtonian Methods ..... 51
4.1 Preliminaries ..... 51
4.2 Forcing Global Convergence ..... 52
4.3 Alleviating the Method ..... 53
4.4 Quasi-Newton Methods ..... 54
4.5 Global Convergence ..... 57
4.6 Local Convergence: Generalities ..... 59
4.7 Local Convergence: BFGS ..... 61
Bibliographical Comments ..... 65
5 Conjugate Gradient ..... 67
5.1 Outline of Conjugate Gradient ..... 67
5.2 Developing the Method ..... 69
5.3 Computing the Direction ..... 70
5.4 The Algorithm Seen as an Orthogonalization Process ..... 70
5.5 Application to Non-Quadratic Functions ..... 72
5.6 Relation with Quasi-Newton ..... 74
Bibliographical Comments ..... 75
6 Special Methods ..... 77
6.1 Trust-Regions ..... 77
6.1.1 The Elementary Problem ..... 78
6.1.2 The Elementary Mechanism: Curvilinear Search ..... 79
6.1.3 Incidence on the Sequence $x_{k}$ ..... 81
6.2 Least-Squares Problems: Gauss-Newton ..... 82
6.3 Large-Scale Problems: Limited-Memory Quasi-Newton ..... 84
6.4 Truncated Newton ..... 86
6.5 Quadratic Programming ..... 88
6.5.1 The basic mechanism ..... 89
6.5.2 The solution algorithm ..... 90
6.5.3 Convergence ..... 92
Bibliographical Comments ..... 95
7 A Case Study: Seismic Reflection Tomography ..... 97
7.1 Modelling ..... 97
7.2 Computation of the Reflection Points ..... 99
7.3 Gradient of the Traveltime. ..... 100
7.4 The Least-Squares Problem to Solve ..... 101
7.5 Solving the Seismic Reflection Tomography Problem ..... 102
General Conclusion ..... 103
Part II Nonsmooth Optimization
8 Introduction to Nonsmooth Optimization ..... 109
8.1 First Elements of Convex Analysis ..... 109
8.2 Lagrangian Relaxation and Duality ..... 111
8.2.1 Primal-Dual Relations ..... 111
8.2.2 Back to the Primal. Recovering Primal Solutions ..... 113
8.3 Two Convex Nondifferentiable Functions ..... 116
8.3.1 Finite Minimax Problems ..... 116
8.3.2 Dual Functions in Lagrangian Duality ..... 117
9 Some Methods in Nonsmooth Optimization ..... 119
9.1 Why Special Methods? ..... 119
9.2 Descent Methods ..... 120
9.2.1 Steepest-Descent Method ..... 121
9.2.2 Stabilization. A Dual Approach. The $\varepsilon$-subdifferential ..... 124
9.3 Two Black-Box Methods ..... 126
9.3.1 Subgradient Methods ..... 127
9.3.2 Cutting-Planes Method ..... 130
10 Bundle Methods. The Quest for Descent ..... 137
10.1 Stabilization. A Primal Approach ..... 137
10.2 Some Examples of Stabilized Problems ..... 140
10.3 Penalized Bundle Methods ..... 141
10.3.1 A Trip to the Dual Space ..... 144
10.3.2 Managing the Bundle. Aggregation ..... 147
10.3.3 Updating the Penalization Parameter. Reversal Forms ..... 150
10.3.4 Convergence Analysis ..... 154
11 Applications of Nonsmooth Optimization ..... 161
11.1 Divide to conquer. Decomposition methods ..... 161
11.1.1 Price Decomposition ..... 163
11.1.2 Resource Decomposition ..... 167
11.1.3 Variable Partitioning or Benders Decomposition ..... 169
11.1.4 Other Decomposition Methods ..... 171
11.2 Transpassing Frontiers ..... 172
11.2.1 Dynamic Bundle Methods ..... 173
11.2.2 Constrained Bundle Methods ..... 177
11.2.3 Bundle Methods for Generalized Equations ..... 180
12 Computational Exercises ..... 183
12.1 Building Prototypical NSO Black Boxes ..... 183
12.1.1 The Function maxQuad ..... 183
12.1.2 The Function maxanal ..... 184
12.2 Implementation of Some NSO Methods ..... 185
12.3 Running the Codes ..... 186
12.4 Improving the Bundle Implementation ..... 187
12.5 Decomposition Application ..... 187
Part III Newton's Methods in Constrained Optimization
13 Background ..... 197
13.1 Differential Calculus ..... 197
13.2 Existence and Uniqueness of Solutions ..... 199
13.3 First-Order Optimality Conditions ..... 200
13.4 Second-Order Optimality Conditions ..... 202
13.5 Speed of Convergence ..... 203
13.6 Projection onto a Closed Convex Set ..... 205
13.7 The Newton Method ..... 205
13.8 The Hanging Chain Project I ..... 208
Notes ..... 213
Exercises ..... 214
14 Local Methods for Problems with Equality Constraints ..... 215
14.1 Newton's Method ..... 216
14.2 Adapted Decompositions of $\mathbb{R}^{n}$ ..... 222
14.3 Local Analysis of Newton's Method ..... 227
14.4 Computation of the Newton Step ..... 230
14.5 Reduced Hessian Algorithm ..... 235
14.6 A Comparison of the Algorithms ..... 243
14.7 The Hanging Chain Project II ..... 245
Notes ..... 250
Exercises ..... 251
15 Local Methods for Problems with Equality and Inequality Constraints ..... 255
15.1 The SQP Algorithm ..... 256
15.2 Primal-Dual Quadratic Convergence ..... 259
15.3 Primal Superlinear Convergence ..... 264
15.4 The Hanging Chain Project III ..... 267
Notes ..... 270
Exercise ..... 270
16 Exact Penalization ..... 271
16.1 Overview ..... 271
16.2 The Lagrangian ..... 274
16.3 The Augmented Lagrangian ..... 275
16.4 Nondifferentiable Augmented Function ..... 279
Notes ..... 284
Exercises ..... 285
17 Globalization by Line-Search ..... 289
17.1 Line-Search SQP Algorithms ..... 291
17.2 Truncated SQP ..... 298
17.3 From Global to Local ..... 307
17.4 The Hanging Chain Project IV ..... 316
Notes ..... 320
Exercises ..... 321
18 Quasi-Newton Versions ..... 323
18.1 Principles ..... 323
18.2 Quasi-Newton SQP ..... 327
18.3 Reduced Quasi-Newton Algorithm ..... 331
18.4 The Hanging Chain Project V ..... 340
Part IV Interior-Point Algorithms for Linear and Quadratic Optimization
19 Linearly Constrained Optimization and Simplex Algorithm ..... 353
19.1 Existence of Solutions ..... 353
19.1.1 Existence Result ..... 353
19.1.2 Basic Points and Extensions ..... 355
19.2 Duality ..... 356
19.2.1 Introducing the Dual Problem ..... 357
19.2.2 Concept of Saddle-Point ..... 358
19.2.3 Other Formulations ..... 362
19.2.4 Strict Complementarity ..... 363
19.3 The Simplex Algorithm ..... 364
19.3.1 Computing the Descent Direction ..... 364
19.3.2 Stating the algorithm ..... 365
19.3.3 Dual simplex ..... 367
19.4 Comments ..... 368
20 Linear Monotone Complementarity and Associated Vector Fields ..... 371
20.1 Logarithmic Penalty and Central Path ..... 371
20.1.1 Logarithmic Penalty ..... 371
20.1.2 Central Path. ..... 372
20.2 Linear Monotone Complementarity ..... 373
20.2.1 General Framework ..... 374
20.2.2 A Group of Transformations ..... 377
20.2.3 Standard Form ..... 378
20.2.4 Partition of Variables and Canonical Form ..... 379
20.2.5 Magnitudes in a Neighborhood of the Central Path ..... 380
20.3 Vector Fields Associated with the Central Path ..... 382
20.3.1 General Framework ..... 383
20.3.2 Scaling the Problem ..... 383
20.3.3 Analysis of the Directions ..... 384
20.3.4 Modified Field ..... 387
20.4 Continuous Trajectories ..... 389
20.4.1 Limit Points of Continuous Trajectories ..... 389
20.4.2 Developing Affine Trajectories and Directions ..... 391
20.4.3 Mizuno's Lemma ..... 393
20.5 Comments ..... 393
21 Predictor-Corrector Algorithms ..... 395
21.1 Overview ..... 395
21.2 Statement of the Methods ..... 396
21.2.1 General Framework for Primal-Dual Algorithms ..... 396
21.2.2 Weighting After Displacement ..... 397
21.2.3 The Predictor-Corrector Method ..... 397
21.3 A small-Neighborhood Algorithm ..... 398
21.3.1 Statement of the Algorithm. Main Result ..... 398
21.3.2 Analysis of the Centralization Move ..... 398
21.3.3 Analysis of the Affine Step and Global Convergence ..... 399
21.3.4 Asymptotic Speed of Convergence ..... 401
21.4 A Predictor-Corrector Algorithm with Modified Field ..... 402
21.4.1 Principle ..... 402
21.4.2 Statement of the Algorithm. Main Result ..... 404
21.4.3 Complexity Analysis ..... 404
21.4.4 Asymptotic Analysis ..... 405
21.5 A Large-Neighborhood Algorithm ..... 406
21.5.1 Statement of the Algorithm. Main Result ..... 406
21.5.2 Analysis of the Centering Step ..... 407
21.5.3 Analysis of the Affine Step ..... 408
21.5.4 Asymptotic Convergence ..... 408
21.6 Practical Aspects ..... 408
21.7 Comments ..... 409
22 Non-Feasible Algorithms ..... 411
22.1 Overview ..... 411
22.2 Principle of the Non-Feasible Path Following ..... 411
22.2.1 Non-Feasible Central Path ..... 411
22.2.2 Directions of Move ..... 412
22.2.3 Orders of Magnitude of Approximately Centered Points ..... 413
22.2.4 Analysis of Directions ..... 415
22.2.5 Modified Field ..... 418
22.3 Non-Feasible Predictor-Corrector Algorithm ..... 419
22.3.1 Complexity Analysis ..... 420
22.3.2 Asymptotic Analysis ..... 422
22.4 Comments ..... 422
23 Self-Duality ..... 425
23.1 Overview ..... 425
23.2 Linear Problems with Inequality Constraints ..... 425
23.2.1 A Family of Self-Dual Linear Problems ..... 425
23.2.2 Embedding in a Self-Dual Problem ..... 427
23.3 Linear Problems in Standard Form ..... 429
23.3.1 The Associated Self-Dual Homogeneous System ..... 429
23.3.2 Embedding in a Feasible Self-Dual Problem ..... 430
23.4 Practical Aspects ..... 431
23.5 Extension to Linear Monotone Complementarity Problems ..... 433
23.6 Comments ..... 434
24 One-Step Methods ..... 435
24.1 Overview ..... 435
24.2 The Largest-Step Sethod ..... 436
24.2.1 Largest-Step Algorithm ..... 436
24.2.2 Largest-Step Algorithm with Safeguard ..... 436
24.3 Centralization in the Space of Large Variables ..... 437
24.3.1 One-Sided Distance ..... 437
24.3.2 Convergence with Strict Complementarity ..... 441
24.3.3 Convergence without Strict Complementarity ..... 443
24.3.4 Relative Distance in the Space of Large Variables ..... 444
24.4 Convergence Analysis ..... 445
24.4.1 Global Convergence of the Largest-Step Algorithm ..... 445
24.4.2 Local Convergence of the Largest-Step Algorithm ..... 446
24.4.3 Convergence of the Largest-Step Algorithm with Safeguard ..... 447
24.5 Comments ..... 450
25 Complexity of Linear Optimization Problems with Integer Data ..... 451
25.1 Overview ..... 451
25.2 Main Results ..... 452
25.2.1 General Hypotheses ..... 452
25.2.2 Statement of the Results ..... 452
25.2.3 Application ..... 453
25.3 Solving a System of Linear Equations ..... 453
25.4 Proofs of the Main Results ..... 455
25.4.1 Proof of Theorem 25.1 ..... 455
25.4.2 Proof of Theorem 25.2 ..... 455
25.5 Comments ..... 456
26 Karmarkar's Algorithm ..... 457
26.1 Overview ..... 457
26.2 Linear Problem in Projective Form ..... 457
26.2.1 Projective Form and Karmarkar Potential ..... 457
26.2.2 Minimizing the Potential and Solving (PLP) ..... 458
26.3 Statement of Karmarkar's Algorithm ..... 459
26.4 Analysis of the Algorithm ..... 460
26.4.1 Complexity Analysis ..... 460
26.4.2 Analysis of the Potential Decrease ..... 460
26.4.3 Estimating the Optimal Cost ..... 461
26.4.4 Practical Aspects ..... 462
26.5 Comments ..... 463
References ..... 465
Index ..... 485

## Part III

## Newton's Methods in Constrained Optimization

J. Charles Gilbert

In this part, we introduce and study numerical techniques based on Newton's method to solve nonlinear optimization problems: objective function and functional constraints can all be nonlinear, possibly nonconvex. Such methods, in the form called sequential quadratic programming (SQP), date back at least to R.B. Wilson's thesis in 1963 [359], but were mainly popularized in the mid-seventies with the appearance of their quasi-Newton versions and their globalization, see U.M. Garcia Palomares and O.L. Mangasarian [280], S.P. Han [184, 185], M.J.D. Powell [291, 292, 293], and the references therein; let us also mention the earlier contributions by B.N. Pshenichnyj [300] and S.M. Robinson [306, 307]. Ongoing research on SQP deals with the efficient use of second derivatives, particularly for nonconvex or large-scale problems, the use of trust regions [86], the treatment of singular or nearly singular situations and of equilibrium constraints [242], globalization by filters, etc. SQP also appears as an auxiliary tool in interior point methods for nonlinear programming [65].

Like Newton's algorithm in unconstrained optimization, SQP is more a methodology than a single algorithm. Here, the basic idea is to linearize the optimality conditions of the problem and to express the resulting linear system in a form suitable for calculation. The interest of linearization is that it provides algorithms with fast local convergence. The linear system is made up of equalities and inequalities, and is viewed as the optimality conditions of a quadratic program. Thus, SQP transforms a nonlinear optimization problem into a sequence of quadratic optimization problems (quadratic objective, linear equality and inequality constraints), which are simpler to solve. This process justifies the name of the SQP family of algorithms. The approach is attractive because efficient algorithms are available to solve quadratic problems: active-set methods [160, 128], augmented Lagrangian techniques [98], and interior-point methods (for the last, see part IV of the present volume).

The above-mentioned principle alone is not sufficient to derive an implementable algorithm. In fact, one must specify how to solve the quadratic program, how to deal with its possible inconsistency, how to cope with a first iterate that is far from a solution (globalization of the method), how the method can be used without computing second derivatives (quasi-Newton versions), how to take advantage of the negative curvature directions, etc. These questions have several answers, whose combinations result in various algorithms, more or less adapted to a particular situation. There is little to be gained from our describing each of these algorithms. Rather, our aim is to present the concepts that form the building blocks of these methods and to show why they are relevant. A good understanding of these tools should allow the reader to adapt the algorithm to a particular problem or to choose the right options of a solver, in order to make it more efficient.

The present review of Newton-like methods for constrained optimization is probably more analysis- than practice-oriented. The aim in this short account is to make an inventory of the main techniques that are continuously
used to analyze these algorithms. In particular, we state and prove precise results on their properties. We also introduce and explain their structure in some detail. However, theory does not cover all aspects of an algorithm. We therefore strive to describe some heuristics that are important for efficient implementations. In fact, it is no exaggeration to say that a method is primarily judged good on the basis of its numerical efficiency. The analysis often comes afterwards to try to explain such a good behavior. Finally, let us mention that all the mathematical concepts used in the present text are simple. In particular, even though we use nonsmooth merit functions, very few notions of nonsmooth analysis are employed, so as to make the text accessible to many.

This part is organized as follows. We start in chapter 13 by recalling some theory on constrained optimization (optimality conditions, constraint qualification, projection onto a convex set, etc.) and Newton's method for nonlinear equations and unconstrained minimization. This chapter ends with the presentation of a numerical project that will go with us along the next chapters of this part (in $\S \S 14.7,15.4,17.4$, and 18.4 ). This project will give us the opportunity to discuss fine points of the implementation of some of the proposed algorithms and to illustrate their behavior in various situations; it also shows, incidentally, that it is relatively easy to write one's own SQP code, provided a solver of quadratic optimization problems is available.

After these preliminaries come two chapters dealing with local methods, whose convergence is ensured if the first iterate is sufficiently close to a solution. Chapter 14 is devoted to problems with only equality constraints. Here we are in the familiar domain of Analysis, where the objects involved (functions and feasible sets) are smooth. The tools are classical as well: mainly linear algebra and differential calculus. A few concepts of differential geometry may be useful to interpret the algorithms. Chapter 15 considers the case where equalities and inequalities are present. Introducing inequality constraints results in an important additional difficulty, due to intrinsic combinatorics in the problem. This comes from the fact that one does not know a priori which inequality constraints are active at a solution, i.e., those that vanish at a solution. If they were known, the algorithms from chapter 14 would apply. The algorithms themselves must therefore determine the set of active constraints, among $2^{m_{I}}$ possibilities ( $m_{I}$ being the number of inequality constraints). Combinatorics is a serious difficulty for algorithms, but SQP copes with it by gracefully forwarding it to a quadratic subproblem, where it is easier to manage. This also implies a change of style in the analysis of the problem. Indeed, various sets of indices must be considered (active or inactive, weakly or strongly active), with an accuracy that is not obtained immediately.

The concept of exact penalty is central to force convergence of algorithms, independently of the initial iterate (a concept known as "globalization"); this is studied in chapter 16. First, the exactness properties of the Lagrangian and
augmented Lagrangian can be analyzed thanks to their smoothness. These results are then used to obtain the exactness of a nondifferentiable merit function. In chapter 17, it is shown how this latter function can be used and how the local algorithms can be modified to obtain convergence of the generated iterates from a starting point that can be far from a solution. The transition from globally convergent algorithms to algorithms with rapid local convergence is also studied in that chapter.

In the quasi-Newton versions of the algorithms, the matrices containing second derivatives are replaced by matrices updated with adequate formulae; this is the subject of chapter 18.

## The Problem to Solve

This text presents efficient algorithms for minimizing a real-valued function $f: \Omega \rightarrow \mathbb{R}$, defined on an open set $\Omega$ in $\mathbb{R}^{n}$, in the presence of functional constraints on the parameters $x=\left(x_{1}, \ldots, x_{n}\right)$ to optimize. Equality constraints $c_{i}(x)=0$, for $i \in E$, as well as inequality constraints $c_{i}(x) \leq 0$, for $i \in I$, can be present. It is supposed that the index sets $E$ (for equalities) and $I$ (for inequalities) are finite, having respectively $m_{E}$ and $m_{I}$ elements. These constraints can also be written

$$
c_{E}(x)=0 \quad \text { and } \quad c_{I}(x) \leq 0 .
$$

Vector inequalities, such as $c_{I}(x) \leq 0$ above, are to be understood componentwise. Hence $c_{I}(x) \leq 0$ means that all the components of the vector $c_{I}(x) \in \mathbb{R}^{m_{I}}$ must be nonpositive. The functions $f$ and $c$ need not be convex.

We therefore look for a point $x_{*} \in \Omega$ that minimizes $f$ on the feasible set

$$
X=\left\{x \in \Omega: c_{E}(x)=0, c_{I}(x) \leq 0\right\} .
$$

A point in $X$ is said to be feasible. The problem is written in a condensed way as follows:

$$
\left(P_{E I}\right) \quad\left\{\begin{array}{l}
\min _{x} f(x) \\
c_{E}(x)=0 \\
c_{I}(x) \leq 0 \\
x \in \Omega
\end{array}\right.
$$

The open set $\Omega$ appearing in $\left(P_{E I}\right)$ cannot be used to express general constraints, since a solution cannot belong to its boundary. It is simply the domain of definition of the functions $f, c_{E}$, and $c_{I}$. It is also the set where some useful properties are satisfied. For example, we always suppose that $c_{E}$ is a submersion on $\Omega$, i.e., that its Jacobian matrix at $x \in \Omega$,

$$
A_{E}(x):=\nabla c_{E}(x)^{\top}
$$

of dimension $m_{E} \times n$ (the rows of $A_{E}(x)$ contain the transposed gradients $\nabla c_{i}(x)^{\top}, i \in E$, for the Euclidean scalar product), is surjective (or onto), for
any $x \in \Omega$. Also, $f$ and $c$ are assumed to be smooth on $\Omega$, for example of class $C^{2}$ (twice continuously differentiable).

We recall from definition 2.2 that problem $\left(P_{E I}\right)$ is said to be convex when $\Omega$ is convex, $f$ and the components of $c_{I}$ are convex and $c_{E}$ is affine. In this case, the feasible set $X$ is convex.

## Notation

We denote by

$$
m=m_{E}+m_{I}
$$

the total number of functional constraints. It will be often convenient to assume that $E$ and $I$ form a partition of $\{1, \ldots, m\}$ :

$$
E \cup I=\{1, \ldots, m\} \quad \text { and } \quad E \cap I=\emptyset
$$

Then, for $v \in \mathbb{R}^{m}$, we denote by $v_{E}$ the $m_{E}$-uple made up of the components $v_{i}$ of $v$, with indices $i \in E$; likewise for $v_{I}$. The constraints $c_{E}$ and $c_{I}$ are then considered to be obtained from a single function $c: \Omega \rightarrow \mathbb{R}^{m}$, whose components indexed in $E$ [resp. $I$ ] form $c_{E}$ [resp. $\left.c_{I}\right]$.

With a vector $v \in \mathbb{R}^{m}$, one associates the vector $v^{\#} \in \mathbb{R}^{m}$, defined as follows:

$$
\left(v^{\#}\right)_{i}=\left\{\begin{array}{l}
v_{i} \text { if } i \in E \\
v_{i}^{+} \text {if } i \in I
\end{array}\right.
$$

where $v_{i}^{+}=\max \left(0, v_{i}\right)$. With this notation, $\left(P_{E I}\right)$ is concisely written as:

$$
\left\{\begin{array}{l}
\min _{x} f(x) \\
c(x)^{\#}=0 \\
x \in \Omega
\end{array}\right.
$$

Indeed, $c(x)^{\#}=0$ if and only if $c_{E}(x)=0$ and $c_{I}(x) \leq 0$.
Let $x \in \Omega$. If $c_{i}(x)=0$, the constraint $i$ is said to be active at $x$. We denote by

$$
I^{0}(x)=\left\{i \in I: c_{i}(x)=0\right\}
$$

the set of indices of inequality constraints that are active at $x \in \Omega$.
The Euclidean or $\ell_{2}$ norm is denoted by $\|\cdot\|_{2}$. We use the same notation for the associated matrix norm.

## Codes

A number of pieces of software based on the algorithmic techniques presented in this part have been written. We give a few words on some of them with a vocabulary that will be clear only after having read part III of the book.

- VF02AD by Powell [293; 1978] is part of the Harwell library. It uses Fletcher's VE02AD code (also part of the Harwell library) for solving the osculating quadratic problems [125].
- NLPQL by Schittkowski [323; 1985-86] can be found in the IMSL library. The osculating quadratic problems are solved by the dual method of Goldfarb and Idnani [165] with the modification proposed by Powell [296] (QL code).
- NPSOL by Gill, Murray, Saunders, and Wright [158; 1986] is available in the NAG library.
- FSQP by Lawrence, Tits, and Zhou [282, 222, 223, 224; 1993-2001] uses an SQP algorithm that evaluates the objective function only at points satisfying the inequality constraints. This nice property can be important for certain classes of applications.
- SPRNLP by Betts and Frank [29; 1994] can use second derivatives (if not positive definite, the Hessian of the Lagrangian is modified using a Levenberg parameter) and exploits sparsity information. It has been used to solve many optimal control problems after a direct transcription discretization.
- FAIPA by Herskovits et al. [189, 190; 1995-1998] also forces the iterates to be strictly feasible with respect to the inequality constraints. Interestingly, the algorithm requires to solve only linear systems of equations, no quadratic optimization problems [283]. This approach is connected to interior point algorithms.
- DONLP2 by Spellucci $[342 ; 1998]$ is available on Netlib. It uses an active set technique on the nonlinear problem, so that the osculating quadratic problems have only equality constraints.
- SNOPT by Gill, Murray, and Saunders [156; 2002] is designed for sparse large-scale problems. The Hessian of the Lagrangian is approximated by limited memory BFGS updates (§6.3). The quadratic programs are solved approximately by an active set method. The globalization is done by linesearch on an augmented Lagrangian merit function.
- SQPAL by Delbos, Gilbert, Glowinski, and Sinoquet [99; 2006] can solve large-scale problems since it uses an augmented Lagrangian approach for solving the quadratic problems [98], a method that has the property of identifying the active constraints in a finite number of iterations.


## Notes

Surveys on Newton's method for constrained optimization have been written by Bertsekas [26; 1982], Powell [295; 1986], Fletcher [128; 1987], Gill, Murray, Saunders, and Wright [159; 1989], Spellucci [340; 1993], Boggs and Tolle [35;

1995], Polak [285; 1997], Sargent [320; 1997], Nocedal and Wright [277; 1999, Chapter 18], Conn, Gould, and Toint [86; 2000, Chapter 15], and Gould, Orban, and Toint [178; 2005]. See also [242] for problems with equilibrium constraints and [28, 325] for applications to optimal control problems.

## Acknowledgement

This part of the book has benefited from the remarks and constructive comments by friends and colleagues, including Paul Armand, Laurent Chauvier, Frédéric Delbos, Sophie Jan-Jégou, Xavier Jonsson, Jean Roberts, Delphine Sinoquet, and the other authors of this book. Some students of ENSTA (École Nationale de Techniques Avancées, Paris) and DEA students of Paris I Panthéon-Sorbonne, to whom the subject of this part has been taught for several years, have also contributed to improve the accessibility of this text. Finally, I would like to thank Richard James, who kindly accepted to supervise the English phrasing of several chapters. His careful reading and erudite recommendations were greatly appreciated.

## 14 Local Methods for Problems with Equality Constraints

In this chapter, we present and study several local methods for minimizing a nonlinear function subject only to nonlinear equality constraints. This is the problem $\left(P_{E}\right)$ represented in figure 14.1: $\Omega$ is an open set of $\mathbb{R}^{n}$, while


Fig. 14.1. Problem $\left(P_{E}\right)$ and its feasible set
$f: \Omega \rightarrow \mathbb{R}$ and $c: \Omega \rightarrow \mathbb{R}^{m}$ are differentiable functions. Since we always assume that $c$ is a submersion, which means that $c^{\prime}(x)$ is surjective (or onto) for all $x \in \Omega$, the inequality $m<n$ is natural. Indeed, for the Jacobian of the constraints to be surjective, we must have $m \leq n$; but if $m=n$, any feasible point is isolated, which results in a completely different problem, for which the algorithms presented here are hardly appropriate. Therefore, a good geometrical representation of the feasible set of problem $\left(P_{E}\right)$ is that of a submanifold $\mathcal{M}_{*}$ of $\mathbb{R}^{n}$, like the one depicted in figure 14.1.

There are several reasons for postponing the study of optimization problems with inequality constraints. First, we tackle difficulties and notation progressively, and prepare the intuition for the general case. Also, the reduced Hessian method (§14.5) has no simple equivalent form when inequalities are present. Finally, such problems arise both in their own right and as subproblems in some algorithmic approaches to solve optimization problems with inequality constraints. For instance, nonlinear interior point algorithms sometimes transform an inequality constrained problem into a sequence a equality constrained problems by introducing slack or shift variables and a logarithmic penalization (see $[143,65,11]$ for examples). A good mastery of the techniques used to solve problem $\left(P_{E}\right)$ is therefore helpful.

By local methods, we mean methods whose convergence is ensured provided the initial iterate is close enough to a solution. In this case, the algorithms presented in chapters 14 and 15 have the nice property to converge quadratically. This feature comes from the linearization of the optimality conditions. Among the quadratically convergent algorithms that have been proposed to solve problem $\left(P_{E}\right)$, we have chosen to describe two of them (and some of their useful variants): Newton's method (§14.1) and the reduced Hessian method (§14.5). These are probably the most often implemented algorithms. Also, they offer a framework in which different techniques can be used: line-search and trust region globalization techniques, quasi-Newton Hessian approximations, etc.

When $c$ is a submersion, the feasible set of $\left(P_{E}\right)$ forms a submanifold of $\mathbb{R}^{n}$. However, the algorithms studied in this section do not force the iterates to stay in that manifold. For general nonlinear constraints, this would generally require too much computing time. Rather, optimality and feasibility are searched simultaneously, so that optimality is obtained in a time of the same order of magnitude as that needed to obtain feasibility in a code without optimization. This nice feature makes these algorithms very attractive in practice.

According to the first-order optimality conditions (13.1), we know that, when the constraints are qualified at a solution $x_{*} \in \Omega$ to $\left(P_{E}\right)$, there exists a Lagrange multiplier $\lambda_{*} \in \mathbb{R}^{m}$ such that

$$
\left\{\begin{array}{l}
\nabla f\left(x_{*}\right)+A\left(x_{*}\right)^{\top} \lambda_{*}=0  \tag{14.1}\\
c\left(x_{*}\right)=0 .
\end{array}\right.
$$

We have denoted by $A(x):=c^{\prime}(x)$ the $m \times n$ Jacobian matrix of the constraints: the $i$ th row of $A(x)$ is the transposed gradient $\nabla c_{i}(x)^{\top}$ of the $i$ th constraint; hence the $(i, j)$ th element of $A(x)$ is the partial derivative $\partial c_{i} / \partial x_{j}(x)$.

### 14.1 Newton's Method

## The Newton Step

We have seen in chapter 13 how Newton's method can be used to solve nonlinear equations (see (13.18)) and to minimize a function (see (13.24)). For optimization problems with equality constraints, it is therefore tempting to compute the step $d_{k}$ at $x_{k}$ by means of a quadratic [resp. linear] approximation of the objective function [resp. constraints] at $x_{k}$. With such a method, $d_{k}$ would solve or would compute a stationary point of the quadratic problem

$$
\left\{\begin{array}{l}
\min _{d} f^{\prime}\left(x_{k}\right) \cdot d+\frac{1}{2} f^{\prime \prime}\left(x_{k}\right) \cdot d^{2}  \tag{14.2}\\
c\left(x_{k}\right)+c^{\prime}\left(x_{k}\right) \cdot d=0
\end{array}\right.
$$

and the next iterate would be $x_{k+1}=x_{k}+d_{k}$. Beware of the nonconvergence of this algorithm! In some cases, the generated sequence moves away from a solution, no matter how close the initial iterate is to this solution ${ }^{1}$.

The right approach consists in dealing simultaneously with the objective minimization and the constraint satisfaction, by working on the optimality conditions (14.1). Actually, these form a system of $n+m$ nonlinear equations in the $n+m$ unknowns $\left(x_{*}, \lambda_{*}\right)$, a system that can be solved by Newton's method. This results is a so-called primal-dual method, which means that a sequence $\left\{\left(x_{k}, \lambda_{k}\right)\right\}$ is generated, in which $x_{k}$ approximates a primal solution $x_{*}$ and $\lambda_{k}$ approximates the associated dual solution $\lambda_{*}$.

Let $\left(x_{k}, \lambda_{k}\right)$ be the current primal-dual iterate. We use the notation
$f_{k}:=f\left(x_{k}\right), \quad c_{k}:=c\left(x_{k}\right), \quad A_{k}:=A\left(x_{k}\right):=c^{\prime}\left(x_{k}\right), \quad \nabla_{x} \ell_{k}:=\nabla_{x} \ell\left(x_{k}, \lambda_{k}\right)$,
and finally denote by

$$
L_{k}:=L\left(x_{k}, \lambda_{k}\right):=\nabla_{x x}^{2} \ell\left(x_{k}, \lambda_{k}\right)
$$

the Hessian of the Lagrangian $\ell$ with respect to $x$ at $\left(x_{k}, \lambda_{k}\right)$. See (13.2) for a definition of the Lagrangian. Newton's method defines a step in $(x, \lambda)$ at $\left(x_{k}, \lambda_{k}\right)$ by linearizing the system (14.1) at $\left(x_{k}, \lambda_{k}\right)$. One finds

$$
\left(\begin{array}{cc}
L_{k} & A_{k}^{\top}  \tag{14.3}\\
A_{k} & 0
\end{array}\right)\binom{d_{k}}{\mu_{k}}=-\binom{\nabla_{x} \ell_{k}}{c_{k}}
$$

Given a solution $\left(d_{k}, \mu_{k}\right)$ to (14.3), the Newton method defines the next iterate $\left(x_{k+1}, \lambda_{k+1}\right)$ by

$$
\begin{equation*}
x_{k+1}=x_{k}+d_{k} \quad \text { and } \quad \lambda_{k+1}=\lambda_{k}+\mu_{k} . \tag{14.4}
\end{equation*}
$$

Since $\nabla_{x} \ell_{k}$ is linear with respect to $\lambda_{k},(14.3)$ can be rewritten as follows:

$$
\left(\begin{array}{cc}
L_{k} & A_{k}^{\top}  \tag{14.5}\\
A_{k} & 0
\end{array}\right)\binom{d_{k}}{\lambda_{k}^{\mathrm{QP}}}=-\binom{\nabla f_{k}}{c_{k}}
$$

where we have used the notation

$$
\lambda_{k}^{\mathrm{QP}}:=\lambda_{k}+\mu_{k}
$$

The superscript 'QP' suggests the fact that, as we shall see below, $\lambda_{k}^{\mathrm{QP}}$ is the multiplier associated with the constraints of a quadratic problem. The next iterate $\left(x_{k+1}, \lambda_{k+1}\right)$ of Newton's method is in this case

[^0]\[

$$
\begin{equation*}
x_{k+1}=x_{k}+d_{k} \quad \text { and } \quad \lambda_{k+1}=\lambda_{k}^{\mathrm{QP}} \tag{14.6}
\end{equation*}
$$

\]

This formulation reveals the less important role played by $\lambda_{k}$, compared with that of $x_{k}$. Observe indeed in (14.5) that $\lambda_{k}$ only appears in the matrix $L_{k}$, while $x_{k}$ is the linearization point of the functions defining the problem.

## Osculating Quadratic Problems

Just as in the unconstrained case, the Newton equation (14.3) can be viewed as the optimality system of a quadratic problem (QP), namely

$$
\left\{\begin{array}{l}
\min _{d} \nabla_{x} \ell_{k}^{\top} d+\frac{1}{2} d^{\top} L_{k} d  \tag{14.7}\\
c_{k}+A_{k} d=0
\end{array}\right.
$$

This one is called the osculating quadratic problem of $\left(P_{E}\right)$ at $\left(x_{k}, \lambda_{k}\right)$. If we consider (14.5) instead of (14.3), we find

$$
\left\{\begin{array}{l}
\min _{d} \nabla f_{k}^{\top} d+\frac{1}{2} d^{\top} L_{k} d  \tag{14.8}\\
c_{k}+A_{k} d=0,
\end{array}\right.
$$

which is another osculating quadratic problem, whose optimality system is (14.5).

The transformations from (14.3) to (14.7) and from (14.5) to (14.8) call for some comments.

1. Any linear system with a symmetric matrix having the structure of that in (14.5) (the distinguishing feature is the zero $(2,2)$ block of the matrix) can be viewed as the first order optimality conditions of the associated QP in (14.8). This point of view can be fruitful when numerical techniques to solve (14.5) are designed.
2. We know that (14.7) and (14.8) have the same primal solutions. This can also be deduced by observing that their objective functions only differ in the term $\lambda_{k}^{\top} A_{k} d$, which is the constant $-\lambda_{k}^{\top} c_{k}$ anywhere on the feasible set. However, these problems have different dual solutions. With (14.7), we obtain the step $\mu_{k}$ to add to the multiplier $\lambda_{k}\left(\lambda_{k+1}=\lambda_{k}+\mu_{k}\right)$, while (14.8) gives directly the new multiplier $\left(\lambda_{k+1}=\lambda_{k}^{Q P}\right)$.
3. One can obtain (14.7) directly from $\left(P_{E}\right)$ : the constraints are linearized at the current point $x_{k}$ and the objective function is a quadratic approximation of the Lagrangian at $\left(x_{k}, \lambda_{k}\right)$ (the constant term $\ell\left(x_{k}, \lambda_{k}\right)$ of this approximation can be added to the objective function of (14.7), without changing the solution).
4. Note the difference between (14.2) and (14.8). The former takes the Hessian of the objective function; the latter uses the Hessian of the Lagrangian. The difference between these two Hessians comes from the constraint curvature (sum of the terms $\left(\lambda_{k}\right)_{i} \nabla^{2} c_{i}\left(x_{k}\right)$ ). In order to have fast
convergence, this curvature must be taken into account. This is all the more important when $f$ is nonconvex.
The validity of (14.7) can be justified a posteriori. Indeed the Lagrangian has a minimum in the subspace tangent to the constraints (if the secondorder sufficient conditions of optimality of theorem 13.4 hold); therefore, it makes sense to minimize the quadratic approximation of this Lagrangian, subject to the linearized constraints. Since the same cannot be said of $f,(14.2)$ appears suspect.
We can also make the following remark. To have a chance of being convergent, an algorithm should at least generate a zero displacement when starting at a solution. We see that this property is not enjoyed by (14.2). In fact, if $x_{k}$ solves $\left(P_{E}\right)$, then $c_{k}=0$ and $\nabla f\left(x_{k}\right)^{\top} d=0$ for all $d \in N\left(A_{k}\right)$; hence (14.2) amounts to minimizing $\frac{1}{2} d^{\top} \nabla f\left(x_{k}\right)^{2} d$ on $N\left(A_{k}\right)$. If the Hessian of $f$ is not positive semi-definite in the space tangent to the constraints, which may well happen, then $d=0$ does not solve (14.2) (unbounded problem). In contrast, (14.3) and (14.5) do enjoy this minimal property, insofar as the matrix appearing in these linear systems is nonsingular (see proposition 14.1 below and the comments that follow definition 14.2).
5. No equivalence holds between (14.5) and (14.8): the minimization problem (14.8) may have a stationary point (hence satisfying (14.5)) but no minimum (unbounded problem). Equivalence does hold between (14.5) and (14.8) - or (14.3) and (14.7) - if $L_{k}$ satisfies

$$
d^{\top} L_{k} d>0, \quad \text { for all nonzero } d \text { in } N\left(A_{k}\right)
$$

In fact, in this case, $d \mapsto \nabla f_{k}^{\top} d+\frac{1}{2} d^{\top} L_{k} d$ is quadratic strictly convex on the affine subspace $\left\{d: c_{k}+A_{k} d=0\right\}$. Therefore (14.8) has a unique solution, which solves the optimality equations (14.5). These equations have no other solution (proposition 14.1).
6. From a numerical point of view, the osculating quadratic problem shows that the Newton equations can be solved by minimization algorithms. For large-scale problems, the reduced conjugate gradient algorithm is often used: one computes a restoration step $r_{k}$ that is feasible for (14.8) (hence satisfying $c_{k}+A_{k} r_{k}=0$ ) and then one generates directions in the null space of $A_{k}$. We shall come back to this issue in $\S 14.4$ and $\S 17.2$.

## Regular Stationary Points

The Newton step can be computed if the linear system that defines it, (14.5) say, is nonsingular. The next proposition gives conditions equivalent to this nonsingularity.

Proposition 14.1 (regular stationary point). Let $A$ be an $m \times n$ matrix, $L$ be an $n \times n$ symmetric matrix, and

$$
K:=\left(\begin{array}{cc}
L & A^{\top}  \tag{14.9}\\
A & 0
\end{array}\right)
$$

Then the following conditions are equivalent:
(i) $K$ is nonsingular;
(ii) $A$ is surjective and any $d \in N(A)$ satisfying $L d \in N(A)^{\perp}$ vanishes;
(iii) $A$ is surjective and $Z^{-\top} L Z^{-}$is nonsingular for some (or any) $n \times$ $(n-m)$ matrix $Z^{-}$whose columns form a basis of $N(A)$.

Proof. $\quad[(i) \Rightarrow(i i)]$ Since $K$ is surjective, so is $A$. On the other hand, if $d \in N(A)$ satisfies $L d \in N(A)^{\perp}=R\left(A^{\top}\right)$, there exists $\mu \in \mathbb{R}^{m}$ such that $(d, \mu) \in N(K)$, so that $d=0$.
$[(i i) \Rightarrow(i i i)]$ Let $Z^{-}$be a matrix like in (iii). If $Z^{-\top} L Z^{-} u=0$ for some $u \in \mathbb{R}^{n-m}, d:=Z^{-} u \in N(A)$ and $L d \in N\left(Z^{-\top}\right)=R\left(Z^{-}\right)^{\perp}=N(A)^{\perp}$, so that $Z^{-} u=0$ by $(i i)$. Now $u=0$ by the injectivity of $Z^{-}$.
$[(i i i) \Rightarrow(i)]$ It suffices to show that $K$ is injective. Take $(d, \mu)$ in its null space. Then $A d=0$ and $L d+A^{\top} \mu=0$, which imply $d \in N(A)$ (or $d=Z^{-} u$ for some $u$ ) and $Z^{-\top} L d=0$. From (iii), $u=0$ and $d=0$. Thus $A^{\top} \mu=0$, and $\mu=0$ by the injectivity of $A^{\top}$.

Note that the nonsingularity of $L$ and the surjectivity of $A$ are not sufficient to guarantee the equivalent conditions $(i)-(i i i)$. For a counter-example consider

$$
L=\left(\begin{array}{cc}
1 & 0 \\
0 & -1
\end{array}\right) \quad \text { and } \quad A=(1-1)
$$

The vector $(11-1)^{\top}$ is in the null space of $K$. On the other hand, when $A$ is surjective, condition (iii) is obviously satisfied if $Z^{-\top} L Z^{-}$is positive definite, and a fortiori if $L$ is positive definite. Exercise 14.2 gives more information on the spectrum of the matrix $K$ : it is claimed in particular that, when $A$ is surjective, the matrix $K$ always has $m$ negative and $m$ positive eigenvalues (for the intuition, consider the case when $n=m=1$ and observe that the determinant of $K$ is negative; hence there is always one negative and one positive eigenvalue).

A consequence of exercise 14.2 is that a quadratic function, whose Hessian is the matrix $K$ with a surjective $A$, is never bounded below. If this function has a stationary point, it is not a minimizer, but a saddle-point. The symmetry of $K$ suggests, however, that a linear system based on this matrix expresses the optimality conditions of a quadratic minimization problem, but this one needs linear equality constraints (using the matrix $A$ ) to have a chance of being well-posed: see (14.8) for an example. Actually, a stationary point of this constrained quadratic problem will be a constrained minimizer if and only if the matrix $L$ is positive semi-definite on the null space of $A$.

The discussion above leads us to introduce the following definition.

Definition 14.2 (regular stationary point). A stationary point ( $x_{*}, \lambda_{*}$ ) of $\left(P_{E}\right)$ is said to be regular if $A_{*}:=c^{\prime}\left(x_{*}\right)$ is surjective and if $Z_{*}^{-\top} L_{*} Z_{*}^{-}$is nonsingular, for some (or any) $n \times(n-m)$ matrix $Z_{*}^{-}$whose columns form a basis of $N\left(A_{*}\right)$.

A regular stationary point is necessarily isolated: it has a neighborhood containing no other stationary point (see exercise 14.3 for a precise statement). Also, a strong primal-dual solution $\left(x_{*}, \lambda_{*}\right)$ to $\left(P_{E}\right)$ satisfying (LI-CQ) (i.e., $A_{*}$ surjective) is a regular stationary point. Indeed, in this case $d^{\top} L_{*} d>0$ for all nonzero $d \in N\left(A_{*}\right)$, so that the so-called reduced Hessian of the Lagrangian

$$
H_{*}:=Z_{*}^{-\top} L_{*} Z_{*}^{-}
$$

is positive definite. The $(n-m) \times(n-m)$ matrix $H_{*}$ clearly depends on the choice of the matrix $Z_{*}^{-}$. In some cases, it can be viewed as a Hessian of some function (see exercise 14.4).

## The Algorithm

We conclude this section by giving a precise description of Newton's algorithm to solve problem $\left(P_{E}\right)$. As already mentioned, the method generates a primaldual sequence $\left\{\left(x_{k}, \lambda_{k}\right)\right\} \subset \mathbb{R}^{n} \times \mathbb{R}^{m}$.

Newton's Algorithm for $\left(P_{E}\right)$ :
Choose an initial iterate $\left(x_{1}, \lambda_{1}\right) \in \mathbb{R}^{n} \times \mathbb{R}^{m}$. Compute $c\left(x_{1}\right), \nabla f\left(x_{1}\right)$, and $A\left(x_{1}\right)$.
Set $k=1$.

1. Stop if $\nabla \ell\left(x_{k}, \lambda_{k}\right)=0$ and $c\left(x_{k}\right)=0$ (optimality is reached).
2. Compute $L\left(x_{k}, \lambda_{k}\right)$ and find a primal-dual stationary point of the quadratic problem (14.8), i.e., a solution $\left(d_{k}, \lambda_{k}^{\mathrm{QP}}\right)$ to (14.5).
3. Set $x_{k+1}:=x_{k}+d_{k}$ and $\lambda_{k+1}:=\lambda_{k}^{\mathrm{QP}}$.
4. Compute $c\left(x_{k+1}\right), \nabla f\left(x_{k+1}\right)$, and $A\left(x_{k+1}\right)$.

5 . Increase $k$ by 1 and go to 1 .

In practice, the stopping criterion in step 1 would test whether $\left\|\nabla \ell\left(x_{k}, \lambda_{k}\right)\right\|$ and $\left\|c\left(x_{k}\right)\right\|$ are sufficiently small. This remark holds for all the algorithms of this part of the book.

Before analyzing the convergence properties of this algorithm in $\S 14.3$, we introduce some notation that makes it easier to understand some interesting variants of the method and highlights the structure of the Newton step $d_{k}$. How to compute this step is dealt with in $\S 14.4$.

### 14.2 Adapted Decompositions of $\mathbb{R}^{n}$

## A General Framework

Suppose that $c$ is a submersion on the open set $\Omega \subset \mathbb{R}^{n}$. Then, the set

$$
\mathcal{M}_{x}:=\{y \in \Omega: c(y)=c(x)\}
$$

is a submanifold of $\mathbb{R}^{n}$ with dimension $n-m$ (for the few concepts of differential geometry that we use, we refer the reader to [344, 51, 84, 112] for example). Intuitively, the tangent space to $\mathcal{M}_{x}$ at $x$ is the set of directions of $\mathbb{R}^{n}$ along which $c$ does not vary at the first order; it is therefore the null space of the Jacobian matrix

$$
A_{x}:=A(x):=c^{\prime}(x)
$$

of $c$ at $x$. This null space and a complementary subspace decompose $\mathbb{R}^{n}$ into two subspaces, which make the description and interpretation of the algorithms easier. This decomposition, which we now describe, is shown in figure 14.2.


Fig. 14.2. Adapted decomposition of $\mathbb{R}^{n}$

Consider first the tangent subspace $N\left(A_{x}\right)$. We shall often assume that we have a smooth mapping

$$
Z^{-}: \Omega \rightarrow \mathbb{R}^{n \times(n-m)}: x \mapsto Z_{x}^{-}:=Z^{-}(x)
$$

such that for all $x \in \Omega, Z_{x}^{-}$is a basis of the tangent subspace. We mean by this that the columns of $Z_{x}^{-}$form a basis of $N\left(A_{x}\right)$ or equivalently:

$$
\begin{equation*}
\forall x \in \Omega, Z_{x}^{-} \text {is } n \times(n-m) \text { injective and } A_{x} Z_{x}^{-}=0 \tag{14.10}
\end{equation*}
$$

Besides, since $A_{x}$ is surjective, it has a right inverse: an $n \times m$ matrix $A_{x}^{-}$ satisfying $A_{x} A_{x}^{-}=I_{m}$. We shall always assume that $A_{x}^{-}$is the value at $x$ of a smooth mapping

$$
A^{-}: \Omega \rightarrow \mathbb{R}^{n \times m}: x \mapsto A_{x}^{-}:=A^{-}(x)
$$

Therefore

$$
\begin{equation*}
\forall x \in \Omega, A_{x}^{-} \text {is } n \times m \text { injective and } A_{x} A_{x}^{-}=I_{m} \tag{14.11}
\end{equation*}
$$

The range space of $A_{x}^{-}$is a subspace complementary to $N\left(A_{x}\right)$, because $R\left(A_{x}^{-}\right) \cap N\left(A_{x}\right)=\{0\}$ and $\operatorname{dim} R\left(A_{x}^{-}\right)+\operatorname{dim} N\left(A_{x}\right)=m+(n-m)=n$.

Thus, $\mathbb{R}^{n}$ can be written as the direct sum of the subspaces spanned by the columns of $Z_{x}^{-}$and the columns of $A_{x}^{-}$: for all $x \in \Omega$,

$$
\mathbb{R}^{n}=R\left(Z_{x}^{-}\right) \oplus R\left(A_{x}^{-}\right) .
$$

Lemma 14.3 (adapted decomposition of $\mathbb{R}^{\boldsymbol{n}}$ ). Let $Z^{-}: \Omega \rightarrow \mathbb{R}^{n \times(n-m)}$ and $A^{-}: \Omega \rightarrow \mathbb{R}^{n \times m}$ be mappings satisfying respectively (14.10) and (14.11). Then there exists a unique mapping

$$
Z: \Omega \rightarrow \mathbb{R}^{(n-m) \times n}: x \mapsto Z_{x}:=Z(x)
$$

satisfying for all $x \in \Omega$ :

$$
\begin{align*}
& Z_{x} A_{x}^{-}=O_{(n-m) \times m}  \tag{14.12}\\
& Z_{x} Z_{x}^{-}=I_{n-m} \tag{14.13}
\end{align*}
$$

This mapping $Z$ is also characterized by the following identity, valid for all $x \in \Omega$ :

$$
\begin{equation*}
I=A_{x}^{-} A_{x}+Z_{x}^{-} Z_{x} \tag{14.14}
\end{equation*}
$$

Proof. It can be easily checked that the matrix $X_{x}=\left(A_{x}^{-} Z_{x}^{-}\right)$is nonsingular, from which follow the existence and uniqueness of $Z_{x}$ satisfying (14.12) and (14.13). Next observe from (14.10), (14.11), (14.12) and (14.13) that the matrix $Y_{x}=\left(A_{x}^{\top} Z_{x}^{\top}\right)^{\top}$ is the inverse of $X_{x}$, since $Y_{x} X_{x}=I_{n}$. Then (14.14) is exactly the identity $X_{x} Y_{x}=I_{n}$. Conversely, this last identity determines $Y_{x}$, hence $Z_{x}$.

Figure 14.2 summarizes the properties of the operators $A_{x}, Z_{x}^{-}, A_{x}^{-}$, and $Z_{x}$. The manifold $\mathcal{M}_{x}$ is translated by $-x$, so that the linearization point $x$ is at the origin. To find one's way in this family of operators, a mnemonic trick is welcome: the operators $A_{x}^{-}$and $Z_{x}^{-}$, with a minus exponent, are injective and right inverses; while the operators $A_{x}$ and $Z_{x}$, without a minus exponent, are surjective and left inverses.

Using the identity (14.14), we have for every vector $v \in \mathbb{R}^{n}$,

$$
v=A_{x}^{-} A_{x} v+Z_{x}^{-} Z_{x} v .
$$

This identity allows us to decompose a vector $v$ into its longitudinal component $Z_{x}^{-} Z_{x} v$, tangent at $x$ to the manifold $\mathcal{M}_{x}$, and its transversal component
$A_{x}^{-} A_{x} v$, which lies in the complementary space $R\left(A_{x}^{-}\right)$. In view of our preceding development, this decomposition is well-defined, once the matrices $Z_{x}^{-}$ and $A_{x}^{-}$have been given. Observe also that $A_{x}^{-} A_{x}$ and $Z_{x}^{-} Z_{x}=I-A_{x}^{-} A_{x}$ are oblique projectors on $R\left(A_{x}^{-}\right)$and $R\left(Z_{x}^{-}\right)$. The orthogonal projectors on these subspaces are

$$
A_{x}^{-}\left(A_{x}^{-\top} A_{x}^{-}\right)^{-1} A_{x}^{-\top}=I-Z_{x}^{\top}\left(Z_{x} Z_{x}^{\top}\right)^{-1} Z_{x}
$$

and

$$
Z_{x}^{-}\left(Z_{x}^{-\top} Z_{x}^{-}\right)^{-1} Z_{x}^{-\top}=I-A_{x}^{\top}\left(A_{x} A_{x}^{\top}\right)^{-1} A_{x} .
$$

Below, we give some formulae for the computation of the matrices $Z_{x}^{-}$and $A_{x}^{-}$satisfying properties (14.10) and (14.11). These formulae use inverses of matrices, which need not be computed explicitly in algorithms. Likewise, the matrices $Z_{x}^{-}$and $A_{x}^{-}$need not be computed explicitly. What matters is their action (or the action of their transpose) on a vector, which can generally be obtained by solving a linear system. For example, as we shall see, the right inverse $A_{x}^{-}$is usually applied to the vector $c(x)$, whereas $A_{x}^{-\top}$ is often applied to $\nabla f(x)$.

We now proceed by giving examples of matrices $Z_{x}^{-}, A_{x}^{-}$, and $Z_{x}$ that are frequently used in the algorithms.

## Decomposition by Partitioning (or Direct Elimination)

This decomposition has its roots in optimal control problems (see §1.2.2 and $\S 1.14$ for examples of such problems), in which the variables $x=(y, u)$ are partitioned in state variables $y \in \mathbb{R}^{m}$ and control variables $u \in \mathbb{R}^{n-m}$. The Jacobian $A_{x}$ is likewise partitioned in

$$
A_{x}=\left(\begin{array}{ll}
B_{x} & N_{x}
\end{array}\right)
$$

where $B_{x}$ is an $m \times m$ matrix giving the derivatives of the constraints with respect to the state variables. In the regular case, $B_{x}$ is nonsingular. Such a decomposition is also used in linear optimization.

The decomposition of $\mathbb{R}^{n}$ given below is often used for large-scale optimization problems, in which a fixed partitioning of the variables leads to a nonsingular matrix $B_{x}$. Note that it is always possible to make a partition of the surjective matrix $A_{x}$ as above, leading to a nonsingular matrix $B_{x}$, provided some permutation of the columns of $A_{x}$ is performed. There are linear solvers that can select the columns of $A_{x}$ in order to form a matrix $B_{x}$ with a reasonably well optimized condition number.

In the framework just described the matrix

$$
\begin{equation*}
Z_{x}^{-}=\binom{-B_{x}^{-1} N_{x}}{I_{n-m}} \tag{14.15}
\end{equation*}
$$

is well defined and satisfies properties (14.10), while the matrix

$$
\begin{equation*}
A_{x}^{-}=\binom{B_{x}^{-1}}{0} \tag{14.16}
\end{equation*}
$$

is also well defined and satisfies (14.11). The mapping $Z$ given by lemma 14.3 has for its value at $x$ :

$$
Z_{x}=\left(O I_{n-m}\right) .
$$

Now, let us highlight some other links with the optimal control framework. Assuming that $c$ is of class $C^{1}$, the nonsingularity of $B_{x}$ implies that $y$, the solution to $c(y, u)=c(x)$ for fixed $x$, is an implicit function of $u: y=y(u)$ and $c(y(u), u)=c(x)$ for all $u$ in a nonempty open set. Then the basis $Z_{x}^{-}$above is obtained by differentiating the parametrization $u \mapsto(y(u), u)$ of the manifold $\mathcal{M}_{x}:=\left\{x^{\prime} \in \Omega: c\left(x^{\prime}\right)=c(x)\right\}$. On the other hand, the displacement

$$
-A_{x}^{-} c(x)=\binom{-B_{x}^{-1} c(x)}{0}
$$

is a Newton step to solve the state equation $c(y, u)=0$, with fixed control $u$.
From a computational point of view, we see that, to evaluate $A_{x}^{-} c(x)$, it is sufficient to solve the linear system $B_{x} v=c(x)$, whose solution $v$ gives the first $m$ components of $A_{x}^{-} c(x)$. This is less expensive than computing $B_{x}^{-1}$ explicitly! Likewise, the first $m$ components $h$ of $Z_{x}^{-} u$ can be obtained by solving the linear system $B_{x} h=-N_{x} u$.

## Orthogonal Decomposition

The orthogonal decomposition is obtained by choosing a right inverse $A_{x}^{-}$, whose columns are perpendicular to $N\left(A_{x}\right)$ (they cannot be orthonormal in general), and a tangent basis $Z_{x}^{-}$with orthonormal columns. The condition on $A_{x}^{-}$implies that this matrix has the form $A_{x}^{-}=A_{x}^{\top} S$, for some matrix $S$. Since $A_{x} A_{x}^{-}=I$ must hold, $A_{x}^{-}$is necessarily given by

$$
\begin{equation*}
A_{x}^{-}=A_{x}^{\top}\left(A_{x} A_{x}^{\top}\right)^{-1} \tag{14.17}
\end{equation*}
$$

Now, let $Z_{x}^{-}$be an arbitrary orthonormal basis of $N\left(A_{x}\right): A_{x} Z_{x}^{-}=0$ and $Z_{x}^{-\top} Z_{x}^{-}=I_{n-m}$. To get the matrix $Z_{x}$ provided by lemma 14.3 , let us multiply both sides of the identity (14.14) to the left by $Z_{x}^{-\top}$, using (14.17). Necessarily

$$
Z_{x}=Z_{x}^{-\top}
$$

One way of computing the matrices $A_{x}^{-}$and $Z_{x}^{-}$just described, is to use the $Q R$ factorization of $A_{x}^{\top}$ (see [170] for example):

$$
A_{x}^{\top}=\left(\begin{array}{ll}
Y_{x}^{-} & Z_{x}^{-} \tag{14.18}
\end{array}\right)\binom{R_{x}}{O}=Y_{x}^{-} R_{x},
$$

where $\left(Y_{x}^{-} Z_{x}^{-}\right)$is an orthogonal matrix and $R_{x}$ is upper triangular. The matrix $R_{x}$ is nonsingular since $A_{x}$ is assumed to be surjective. Then, the
last $n-m$ columns $Z_{x}^{-}$of the orthogonal factor form an orthonormal basis of $R\left(Y_{x}^{-}\right)^{\perp}=R\left(A_{x}^{\top}\right)^{\perp}$, which is indeed the null space of $A_{x}$. Furthermore, (14.18) and the nonsingularity of $R_{x}$ show that the columns of $Y_{x}^{-} \in \mathbb{R}^{n \times m}$ span $R\left(A_{x}^{\top}\right)=N\left(A_{x}\right)^{\perp}$. Since, by multiplying the extreme sides of (14.18) to the left by $Y_{x}^{-\top}$, it follows that $A_{x} Y_{x}^{-}=R_{x}^{\top}$ or $A_{x} Y_{x}^{-} R_{x}^{-\top}=I_{m}$, the right inverse of $A_{x}$ given by (14.17) is necessarily

$$
A_{x}^{-}=Y_{x}^{-} R_{x}^{-\top} .
$$

The orthogonal decomposition just described has the advantage of being numerically stable and of computing a perfectly well-conditioned basis $Z_{x}^{-}$. The $Q R$ factorization can be carried out by using Givens rotations or with at most $m$ Householder reflections. Therefore, this is a viable approach when $m$ is not too large.

## Oblique Decomposition

Let $M$ be a matrix that is nonsingular on the null space of $A_{x}$, meaning that $Z_{x}^{-\top} M Z_{x}^{-}$is nonsingular for some basis $Z_{x}^{-}$of $N\left(A_{x}\right)$ (this property of $M$ does not depend on the choice of $Z_{x}^{-}$, see proposition 14.1). Then, one can associate with $M$ a right inverse of $A_{x}$, defined as follows. Take $v \in \mathbb{R}^{m}$. Then the quadratic problem in $d$

$$
\left\{\begin{array}{l}
\min _{d} \frac{1}{2} d^{\top} M d  \tag{14.19}\\
A_{x} d=v
\end{array}\right.
$$

has a unique stationary point, which satisfies the optimality conditions

$$
\left\{\begin{array}{l}
M d+A_{x}^{\top} \lambda=0  \tag{14.20}\\
A_{x} d=v,
\end{array}\right.
$$

for some multiplier $\lambda \in \mathbb{R}^{m}$. We see that $d$ depends linearly on $v$. Denoting by $\widehat{A}_{x}^{-}$the matrix representing this linear mapping, i.e., $d=\widehat{A}_{x}^{-} v$, the second equation in (14.20) shows that $\widehat{A}_{x}^{-}$is a right inverse of $A_{x}$. This matrix $\widehat{A}_{x}^{-}$ will be useful to write a simple expression of the Newton displacement to solve ( $P_{E}$ ).

An explicit expression of $\widehat{A}_{x}^{-}$can be given by using a basis $Z_{x}^{-}$of the null space of $A_{x}$ and a right inverse $A_{x}^{-}$of $A_{x}$. Then (14.14) and (14.20) show that $d=A_{x}^{-} v+Z_{x}^{-} u$ for some $u \in \mathbb{R}^{n-m}$. By premultiplying both sides of the first equation of (14.20) by $Z_{x}^{-\top}$, we obtain $u=-\left(Z_{x}^{-\top} M Z_{x}^{-}\right)^{-1} Z_{x}^{-\top} M A_{x}^{-} v$. Finally

$$
\begin{equation*}
\widehat{A}_{x}^{-}=\left(I-Z_{x}^{-}\left(Z_{x}^{-\top} M Z_{x}^{-}\right)^{-1} Z_{x}^{-\top} M\right) A_{x}^{-} . \tag{14.21}
\end{equation*}
$$

Even though $A_{x}^{-}$and $Z_{x}^{-}$appear in this formula, $\widehat{A}_{x}^{-}$does not depend on them (from its definition). From lemma 14.3, there corresponds to the operators $Z_{x}^{-}$
and $\widehat{A}_{x}^{-}$a unique matrix $\widehat{Z}_{x}$ such that $\widehat{Z}_{x} \widehat{A}_{x}^{-}=0$ and $\widehat{Z}_{x} Z_{x}^{-}=I$. To give an analytic expression of $\widehat{Z}_{x}$, observe first that from (14.21), one has

$$
\begin{equation*}
Z_{x}^{-\top} M \widehat{A}_{x}^{-}=0 \tag{14.22}
\end{equation*}
$$

which expresses the fact that the range spaces $R\left(Z_{x}^{-}\right)$and $R\left(\widehat{A}_{x}^{-}\right)$are "orthogonal" with respect to the matrix $M$ (this would correspond to a proper notion of orthogonality if the matrix $M$ were positive definite). It is then easy to check that

$$
\widehat{Z}_{x}=\left(Z_{x}^{-\top} M Z_{x}^{-}\right)^{-1} Z_{x}^{-\top} M
$$

satisfies the required properties.
To conclude, note that $\widehat{A}_{x}^{-}$may not exist if $M$ is singular on the null space of $A_{x}$. Here is a counter-example with $n=2$ and $m=1$ :

$$
M=\left(\begin{array}{ll}
1 & 1 \\
1 & 0
\end{array}\right), \quad A_{x}=\left(\begin{array}{ll}
1 & 0
\end{array}\right), \quad \text { and } \quad Z_{x}^{-}=\binom{0}{1}
$$

Since $Z_{x}^{-\top} M=A_{x}, \widehat{A}_{x}^{-}$cannot satisfy both $A_{x} \widehat{A}_{x}^{-}=I$ and (14.22). Observe finally that the right inverses (14.16) and (14.17) obtained previously can be recovered from $\widehat{A}_{x}^{-}$by an appropriate choice of $M$; this is the subject of exercise 14.7.

### 14.3 Local Analysis of Newton's Method

## Local Convergence

In this section, we study the local convergence of the Newton algorithm to solve problem $\left(P_{E}\right)$, introduced in $\S 14.1$. We use the notation

$$
A_{*}=A\left(x_{*}\right) \quad \text { and } \quad L_{*}=L\left(x_{*}, \lambda_{*}\right)
$$

Quadratic convergence of the primal-dual sequence $\left\{\left(x_{k}, \lambda_{k}\right)\right\}$ will be shown thanks to theorem 13.6. We shall also use proposition 14.1, whose conditions (i)-(iii) imply that the constraints are qualified at the solution $x_{*}$ in the sense (LI-CQ):

$$
\begin{equation*}
A_{*} \text { is surjective. } \tag{14.23}
\end{equation*}
$$

A consequence of proposition 14.1 is that, when $\left(x_{*}, \lambda_{*}\right)$ is a regular stationary point, the system (14.3) or (14.5) has a unique solution for $\left(x_{k}, \lambda_{k}\right)$ close to $\left(x_{*}, \lambda_{*}\right)$. Therefore Newton's method is well defined in the neighborhood of regular stationary points.

Theorem 14.4 (convergence of Newton's algorithm). Suppose that $f$ and $c$ are of class $C^{2}$ in a neighborhood of a regular stationary point $x_{*}$ of $\left(P_{E}\right)$, with associated multiplier $\lambda_{*}$. Then, there exists a neighborhood $V$ of
$\left(x_{*}, \lambda_{*}\right)$ such that, if the first iterate $\left(x_{1}, \lambda_{1}\right) \in V$, the Newton algorithm defined in $\S 14.1$ is well-defined and generates a sequence $\left\{\left(x_{k}, \lambda_{k}\right)\right\}$ converging superlinearly to $\left(x_{*}, \lambda_{*}\right)$. If $f^{\prime \prime}$ and $c^{\prime \prime}$ are Lipschitzian in a neighborhood of $x_{*}$, the convergence of the sequence is quadratic.

Proof. The result is obtained by applying theorem 13.6 with $z=(x, \lambda)$ and

$$
F(z)=\binom{\nabla f(x)+A(x)^{\top} \lambda}{c(x)}
$$

Clearly, $F$ is of class $C^{1}$ in a neighborhood of $z_{*}=\left(x_{*}, \lambda_{*}\right)$ and $F^{\prime}\left(z_{*}\right)$ is nonsingular (from proposition 14.1). The superlinear convergence of $\left\{\left(x_{k}, \lambda_{k}\right)\right\}$ to $\left(x_{*}, \lambda_{*}\right)$ follows if $\left(x_{1}, \lambda_{1}\right)$ is close enough to $\left(x_{*}, \lambda_{*}\right)$. If $f^{\prime \prime}$ and $c^{\prime \prime}$ are Lipschitzian near $x_{*}$, so is $F^{\prime}$ near $z_{*}$, and the quadratic convergence of $\left\{\left(x_{k}, \lambda_{k}\right)\right\}$ follows.

This theorem tells us that Newton's algorithm makes no distinction between stationary points, provided they are regular in the sense of definition 14.2. The iterates are indeed attracted by such a point, even if it is not a local minimum of $\left(P_{E}\right)$; in particular it can be a maximum. The reason of this property comes from the fact that Newton's algorithm is essentially a method to solve nonlinear equations (here the optimality conditions of $\left(P_{E}\right)$ ). When one tries to find a minimizer, this is not a nice property. We shall see, however, that the techniques of chapter 17 tends to overcome this undesirable feature.

Note that the quadratic convergence of the sequence $\left\{\left(x_{k}, \lambda_{k}\right)\right\}$ by no means implies that of $\left\{x_{k}\right\}$ (see exercise 14.8). However, we shall see in chapter 15 (theorem 15.7) that $\left\{x_{k}\right\}$ does converge superlinearly. On the other hand, there are versions of Newton's method that guarantee the quadratic convergence of the primal sequence $\left\{x_{k}\right\}$. Here is an example of such an algorithm.

## A Primal Version of the Newton Algorithm

It has already been observed that, in Newton's method, $\lambda_{k}$ plays a less crucial role than $x_{k}$ in the computation of the next iterate $\left(x_{k+1}, \lambda_{k+1}\right)$. If, instead of letting the sequences $\left\{x_{k}\right\}$ and $\left\{\lambda_{k}\right\}$ be generated independently, the dual iterate $\lambda_{k}$ is computed from the primal iterate $x_{k}$, by means of a function $x \mapsto \lambda(x)$, i.e.,

$$
\lambda_{k}=\lambda\left(x_{k}\right),
$$

the algorithm becomes completely primal. Indeed, then the knowledge of $x_{k}$ entirely determines the next iterate $x_{k+1}$. We shall show below that the function $\lambda(\cdot)$ can be chosen in such a way that the convergence of $\left\{x_{k}\right\}$ will be quadratic, under natural assumptions. A possible candidate for that function is the least-squares multiplier:

$$
\begin{equation*}
\lambda^{\mathrm{LS}}(x):=-A^{-}(x)^{\top} \nabla f(x) \tag{14.24}
\end{equation*}
$$

where $A^{-}(x)$ is a right inverse of $A(x)$. One speaks of least-squares multiplier because $\lambda^{\mathrm{LS}}(x)$ minimizes in $\lambda$ a weighted $\ell_{2}$ norm of $\nabla_{x} \ell(x, \lambda)$ (see exercise 14.9).

Let us make precise the algorithm under investigation.

Primal version of Newton's Algorithm for $\left(P_{E}\right)$ :
Choose an initial iterate $x_{1} \in \mathbb{R}^{n}$.
Compute $c\left(x_{1}\right), \nabla f\left(x_{1}\right)$, and $A\left(x_{1}\right)$.
Set $k=1$.

1. Compute $\lambda_{k}=\lambda\left(x_{k}\right)$.
2. Stop if $\nabla \ell\left(x_{k}, \lambda_{k}\right)=0$ and $c\left(x_{k}\right)=0$ (optimality is reached).
3. Compute $L\left(x_{k}, \lambda_{k}\right)$ and find a solution $\left(d_{k}, \lambda_{k}^{\mathrm{QP}}\right)$ to the linear system

$$
\left(\begin{array}{cc}
L\left(x_{k}, \lambda_{k}\right) & A\left(x_{k}\right)^{\top}  \tag{14.25}\\
A\left(x_{k}\right) & 0
\end{array}\right)\binom{d_{k}}{\lambda_{k}^{Q \mathrm{P}}}=-\binom{\nabla f\left(x_{k}\right)}{c\left(x_{k}\right)} .
$$

4. Set $x_{k+1}:=x_{k}+d_{k}$.
5. Compute $c\left(x_{k+1}\right), \nabla f\left(x_{k+1}\right)$, and $A\left(x_{k+1}\right)$.

6 . Increase $k$ by 1 and go to 1 .

We have used the same notation $\lambda_{k}^{Q P}$ for the dual solution to (14.25) and (14.5), although their values are different, since here $\lambda_{k}$ depends on $x_{k}$. Note that although $\lambda_{k}^{\mathrm{QP}}$ is computed, it has no influence on the value of $x_{k+1}$.

The next theorem analyses the local convergence of this algorithm.
Theorem 14.5 (convergence of a primal version of Newton's algorithm). Suppose that $f$ and $c$ are of class $C^{2}$ in a neighborhood of a regular stationary point $x_{*}$ of $\left(P_{E}\right)$, with associated multiplier $\lambda_{*}$. Suppose also that the function $\lambda(\cdot)$ used to set the value of $\lambda_{k}$ satisfies $\lambda\left(x_{*}\right)=\lambda_{*}$ and is continuous at $x_{*}$. Then, there exists a neighborhood $V$ of $x_{*}$ such that, if the first iterate $x_{1} \in V$, the above primal version of Newton's algorithm is well-defined, generates a sequence $\left\{x_{k}\right\}$ converging superlinearly to $x_{*}$, and $\lambda_{k}^{\mathrm{QP}}-\lambda_{*}=o\left(\left\|x_{k}-x_{*}\right\|\right)$. If furthermore $f^{\prime \prime}$ and $c^{\prime \prime}$ are Lipschitzian in a neighborhood of $x_{*}$ and if there is a positive constant $C$ such that

$$
\left\|\lambda(x)-\lambda_{*}\right\| \leq C\left\|x-x_{*}\right\|, \quad \text { for } x \text { near } x_{*},
$$

then the convergence of $\left\{x_{k}\right\}$ is quadratic and $\lambda_{k}^{\mathrm{QP}}-\lambda_{*}=O\left(\left\|x_{k}-x_{*}\right\|^{2}\right)$.
Proof. We mimic the argument used in the proof of theorem 13.6. With the notation

$$
F(x, \nu):=\binom{\nabla_{x} \ell(x, \nu)}{c(x)}
$$

and $\mu_{k}:=\lambda_{k}^{\mathrm{QP}}-\lambda_{*}$, the linear system (14.25) can be written

$$
F^{\prime}\left(x_{k}, \lambda_{k}\right)\binom{d_{k}}{\mu_{k}}=-F\left(x_{k}, \lambda_{*}\right)
$$

If $x_{k}$ is in some neighborhood of the regular stationary point $x_{*}$, with associated multiplier $\lambda_{*}, \lambda_{k}$ is near $\lambda_{*}$ (continuity of $\lambda(\cdot)$ at $x_{*}$ ). Furthermore, $F^{\prime}\left(x_{k}, \lambda_{k}\right)=F^{\prime}\left(x_{k}, \lambda\left(x_{k}\right)\right)$ is nonsingular (see proposition 14.1) and has a bounded inverse on that neighborhood. With the notation $z_{k+1}:=$ $\left(x_{k+1}, \lambda_{k}^{\mathrm{QP}}\right), z_{k, *}:=\left(x_{k}, \lambda_{*}\right)$, and $z_{*}:=\left(x_{*}, \lambda_{*}\right)$, and the fact that $f$ and $c$ are of class $C^{2}$, one has

$$
\begin{aligned}
& z_{k+1}-z_{*}= z_{k, *}-z_{*}-F^{\prime}\left(x_{k}, \lambda_{k}\right)^{-1} F\left(x_{k}, \lambda_{*}\right) \\
&= F^{\prime}\left(x_{k}, \lambda_{k}\right)^{-1}( \\
&\left(F^{\prime}\left(x_{k}, \lambda_{k}\right)\left(z_{k, *}-z_{*}\right)-F\left(z_{*}\right)\right. \\
&\left.-\int_{0}^{1} F^{\prime}\left(x_{*}+t\left(x_{k}-x_{*}\right), \lambda_{*}\right) \cdot\left(z_{k, *}-z_{*}\right) \mathrm{d} t\right) .
\end{aligned}
$$

Using $F\left(z_{*}\right)=0$ and taking norms,

$$
\left\|z_{k+1}-z_{*}\right\| \leq C^{\prime}\left(\int_{0}^{1}\left\|F^{\prime}\left(x_{k}, \lambda_{k}\right)-F^{\prime}\left(x_{*}+t\left(x_{k}-x_{*}\right), \lambda_{*}\right)\right\| \mathrm{d} t\right)\left\|x_{k}-x_{*}\right\|
$$

where $C^{\prime}$ is a positive constant. Now, since $f^{\prime \prime}, c^{\prime \prime}$, and $\lambda$ are continuous at $x_{*}, F^{\prime}(\cdot, \lambda(\cdot))$ is continuous at $x_{*}$ and the last estimate gives $z_{k+1}-z_{*}=$ $o\left(\left\|x_{k}-x_{*}\right\|\right)$, implying the superlinear convergence of $x_{k}$ to $x_{*}$ and $\lambda_{k}^{\mathrm{QP}}-$ $\lambda_{*}=o\left(\left\|x_{k}-x_{*}\right\|\right)$. If furthermore $f^{\prime \prime}$ and $c^{\prime \prime}$ are Lipschitzian near $x_{*}$ and $\lambda(x)-\lambda_{*}=O\left(\left\|x-x_{*}\right\|\right)$, one has $z_{k+1}-z_{*}=O\left(\left\|x_{k}-x_{*}\right\|^{2}\right)$, which means that the convergence of $\left\{x_{k}\right\}$ is now quadratic and that $\lambda_{k}^{\mathrm{QP}}-\lambda_{*}=O\left(\left\|x_{k}-x_{*}\right\|^{2}\right)$.

### 14.4 Computation of the Newton Step

In this section, we describe three ways of computing the Newton step $d_{k}$ and the associated multiplier $\lambda_{k}^{\mathrm{QP}}$ : the direct inversion approach, the dual approach, and the reduced system approach. We are interested both in analytic expressions of $\left(d_{k}, \lambda_{k}^{\mathrm{QP}}\right)$ and computational issues. Each of these methods has its own advantages and drawbacks. It is the last one that most highlights the structure of the Newton step. In each case, one has to find a solution to (14.5), which is recalled here for convenience:

$$
\left(\begin{array}{cc}
L_{k} & A_{k}^{\top}  \tag{14.26}\\
A_{k} & 0
\end{array}\right)\binom{d_{k}}{\lambda_{k}^{\mathrm{QP}}}=-\binom{\nabla f_{k}}{c_{k}} .
$$

Below, the matrix of this linear system is supposed nonsingular (see proposition 14.1 for conditions ensuring this property), which implies that $A_{k}$ is surjective.

## The Direct Inversion Approach

The most straightforward approach for computing the Newton step is to consider the linear system (14.26) as a whole, without exploiting its block structure. One should not lose sight of the dimension $n+m$ of this linear system, which can be quite large in practice. Therefore, to make this approach attractive the problem needs to have small dimensions or to have sparse matrices $L_{k}$ and $A_{k}$ that can be taken into account. Using this approach could also be a naive but rapid way of computing $\left(d_{k}, \lambda_{k}^{\mathrm{QP}}\right)$ in a personal program, using matrix oriented languages like Matlab or Scilab, for instance.

As regards the numerical techniques used to solve the full linear system, observe that, although the matrix in (14.26) is symmetric, it is never positive definite, even at a strong solution to problem $\left(P_{E}\right)$ (see exercise 14.2). Therefore, a Cholesky factorization or conjugate gradient iterations are not adequate algorithms to solve this linear system! Direct linear solvers (i.e., those that factorize the matrix in (14.26)) can be considered, in particular when they can take advantage of the possible sparsity of $A_{k}$ and $L_{k}$. The methods of Bunch and Kaufman [56] for the dense case or the MA27/MA47 solvers of Duff and Reid $[114,115,116]$ for the sparse case are often employed. For large-scale problems, iterative solvers with preconditioners have also been developed, see for example [53, 20, 315, 358, 333].

## The Dual Approaches

The dual approaches (sometimes called range-space approaches) need to have nonsingular matrices $L_{k}$ and $A_{k} L_{k}^{-1} A_{k}^{\top}$. This condition is certainly satisfied if $L_{k}$ is positive definite (remember that $A_{k}$ is always assumed surjective in this section).

In the dual approach, the value of $d_{k}$ is given as a function of $\lambda_{k}^{\mathrm{QP}}$, using the first equation of (14.26):

$$
\begin{equation*}
d_{k}=-L_{k}^{-1}\left(\nabla f_{k}+A_{k}^{\top} \lambda_{k}^{\mathrm{QP}}\right) \tag{14.27}
\end{equation*}
$$

Substituting this expression in the second equation of (14.26) gives the value of the QP multiplier, which is the solution to the linear system

$$
\begin{equation*}
\left(A_{k} L_{k}^{-1} A_{k}^{\top}\right) \lambda_{k}^{\mathrm{QP}}=-A_{k} L_{k}^{-1} \nabla f_{k}+c_{k} . \tag{14.28}
\end{equation*}
$$

A way of solving (14.26) is then to consider the two linear systems (14.28) and (14.27) one after the other: once $\lambda_{k}^{\mathrm{QP}}$ has been determined by (14.28), $d_{k}$ can be evaluated by (14.27). The computational effort depends on how these systems are solved, which should be a consequence of the problem size and structure. If direct solvers are used, one can give a rapid count of the number of linear systems to solve: $m+1$ with the $n \times n$ matrix $L_{k}$ and one with the $m \times m$ matrix $A_{k} L_{k}^{-1} A_{k}^{\top}$. Indeed, the calculation can be organized as follows: first, one computes $L_{k}^{-1} A_{k}^{\top}$ and $L_{k}^{-1} \nabla f_{k}$; next, $\lambda_{k}^{\text {QP }}$ is evaluated
by solving (14.28); finally, $d_{k}$ is obtained by (14.27) without having to solve any additional linear system.

When $L_{k}$ is positive definite, $\lambda_{k}^{\mathrm{QP}}$ in (14.28) maximizes the dual function associated with the osculating quadratic problem (14.8), which is the function (see also part II)

$$
\begin{equation*}
\lambda \mapsto \min _{d}\left(\frac{1}{2} d^{\top} L_{k} d+\nabla f_{k}^{\top} d+\lambda^{\top}\left(c_{k}+A_{k} d\right)\right) \tag{14.29}
\end{equation*}
$$

On the other hand, $d_{k}$ given by (14.27) is the solution to this minimization problem in (14.29) with $\lambda=\lambda_{k}^{Q P}$. This viewpoint gives its name to the approach. It also suggests other ways of solving (14.26), which are often interesting for very large-scale problems such as the Stokes equations in fluid mechanics (see [243] and references therein). We briefly discuss these approaches below.

The Uzawa algorithm $[13,134]$ generates a sequence of multipliers $\lambda$ converging to $\lambda_{k}^{\mathrm{QP}}$. For each $\lambda$, the minimization problem in (14.29) is solved, which provides an approximation $d$ of the solution $d_{k}$. Next the multiplier is updated by a steepest ascent step on the dual function: $\lambda_{+}:=\lambda+\alpha\left(c_{k}+A_{k} d\right)$, where $\alpha>0$ is an "appropriate" stepsize. This first order method in $\lambda$ is sometimes too slow. One way of accelerating it in this simple quadratic setting is to use the conjugate gradient (CG) algorithm on the dual function, which is equivalent to solving the linear system (14.28) by CG. Each CG iteration normally requires an accurate solution to a linear system with the matrix $L_{k}$, although inexact solution can also be considered (see for example [355]).

Another way of accelerating the Uzawa procedure described above is to substitute in (14.29) the Lagrangian by the augmented Lagrangian (see § 16.3):

$$
\begin{equation*}
\lambda \mapsto \min _{d}\left(\frac{1}{2} d^{\top} L_{k} d+\nabla f_{k}^{\top} d+\lambda^{\top}\left(c_{k}+A_{k} d\right)+\frac{r}{2}\left\|c_{k}+A_{k} d\right\|_{2}^{2}\right), \tag{14.30}
\end{equation*}
$$

where $r>0$ is a parameter. The algorithm is similar: $\lambda_{+}:=\lambda+r\left(c_{k}+A_{k} d\right)$, where $d$ is now the solution to the minimization problem in (14.30). See [134] for more details.

Time saving is also possible by avoiding an exact minimization of the problem in (14.29) or (14.30) before updating the multiplier (see [303, 118, 54, 92] for instance).

In conclusion, the dual approaches can be appropriate when $L_{k}$ and $A_{k} L_{k}^{-1} A_{k}^{\top}$ are nonsingular and a linear system with the matrix $L_{k}$ is not too difficult to solve. They can also be useful when quasi-Newton techniques are used to approximate $L_{k}^{-1}$ by positive definite matrices in the nonlinear algorithm (the one that sets problem (14.26)), since then there is no linear system to solve with the matrix $L_{k}$, just a matrix-vector product needs to be done.

## The Reduced System Approach

In this approach (sometimes called the null-space approach), it is assumed that a decomposition of $\mathbb{R}^{n}$ has been chosen, similar to those described in $\S 14.2$. The operators $A^{-}(x)$ and $Z^{-}(x)$ should take advantage of the features of the problem, in order to avoid expensive operations. We show below that then the optimization aspect contained in (14.26) can be transferred into a single linear system, involving an $(n-m) \times(n-m)$ symmetric matrix: the reduced Hessian of the Lagrangian. This makes the reduced system approach particularly appropriate when $n-m \ll n$. Since the reduced Hessian is positive definite at a strong solution to $\left(P_{E}\right)$, the approach makes it possible to detect convergence to a stationary point that is not a local minimum. Furthermore, the method leads to formulae highlighting the structure of the Newton step $d_{k}$.

Let us start by introducing a very useful notion. We have denoted by $Z^{-}(x)$ an $n \times(n-m)$ matrix, whose columns form a basis of $N(A(x))$, the subspace tangent to the constraint manifold at $x$. We call reduced gradient of $f$ at $x$ for the basis $Z^{-}$, the vector of $\mathbb{R}^{n-m}$ defined by

$$
\begin{equation*}
g(x):=Z^{-}(x)^{\top} \nabla f(x) . \tag{14.31}
\end{equation*}
$$

We note $g_{k}:=g\left(x_{k}\right)$. This vector can be interpreted in Riemannian geometry as follows. Equip the manifold $\mathcal{M}_{x}$ with a Riemannian structure by defining at each point $y \in \mathcal{M}_{x}$ the scalar product on the tangent space $\gamma_{y}\left(Z_{y}^{-} u, Z_{y}^{-} v\right)=u^{\top} v$; then the gradient of $\left.f\right|_{\mathcal{M}_{x}}$ at $y$ for this Riemannian metric is just the tangent vector $Z^{-}(y) g(y)$.

Consider now the computation of $d_{k}$. Recalling (14.14), the second equation in (14.26) shows that $d_{k}$ has the form

$$
d_{k}=-A_{k}^{-} c_{k}+Z_{k}^{-} u_{k}
$$

for some $u_{k} \in \mathbb{R}^{n-m}$. Then, the first equation in (14.26) gives

$$
L_{k} Z_{k}^{-} u_{k}+A_{k}^{\top} \lambda_{k}^{\mathrm{QP}}=-\nabla f_{k}+L_{k} A_{k}^{-} c_{k}
$$

Premultiplying by $Z_{k}^{-\top}$ to eliminate $\lambda_{k}^{\mathrm{QP}}$ provides the reduced linear system:

$$
\begin{equation*}
H_{k} u_{k}=-g_{k}+Z_{k}^{-\top} L_{k} A_{k}^{-} c_{k} \tag{14.32}
\end{equation*}
$$

where the $(n-m) \times(n-m)$ matrix

$$
H_{k}:=Z_{k}^{-\top} L_{k} Z_{k}^{-}
$$

is called the reduced Hessian of the Lagrangian at $\left(x_{k}, \lambda_{k}\right)$. It depends on the choice of the basis $Z_{k}^{-}$. This matrix is necessarily nonsingular when the matrix in (14.26) is nonsingular (see proposition 14.1). This leads to

$$
d_{k}=-\left(I-Z_{k}^{-} H_{k}^{-1} Z_{k}^{-\top} L_{k}\right) A_{k}^{-} c_{k}-Z_{k}^{-} H_{k}^{-1} g_{k}
$$

The operator acting on $c_{k}$, namely

$$
\begin{equation*}
\widehat{A}_{k}^{-}:=\left(I-Z_{k}^{-} H_{k}^{-1} Z_{k}^{-\top} L_{k}\right) A_{k}^{-}, \tag{14.33}
\end{equation*}
$$

is the right inverse of $A_{k}$ defined in (14.21), where $M$ and $x$ have been replaced by $L_{k}$ and $x_{k}$ respectively. Finally

$$
\begin{equation*}
d_{k}=-\widehat{A}_{k}^{-} c_{k}-Z_{k}^{-} H_{k}^{-1} g_{k} . \tag{14.34}
\end{equation*}
$$

This computation reveals the structure of the Newton direction $d_{k}$, made up of two terms (see figure 14.3). The first term $\widehat{r}_{k}:=-\widehat{A}_{k}^{-} c_{k}$ is a stationary


Fig. 14.3. Structure of the Newton step $d_{k}$
point of the quadratic problem in $r \in \mathbb{R}^{n}$ :

$$
\left\{\begin{array}{l}
\min _{r} \frac{1}{2} r^{\top} L_{k} r \\
c_{k}+A_{k} r=0
\end{array}\right.
$$

To see this, just set $\nabla f_{k}=0$ in (14.8) and (14.34). This direction aims at reducing $\rho(\cdot)=\|c(\cdot)\|$, an arbitrary norm of the constraints. Indeed, when $c_{k} \neq 0, \widehat{r}_{k}$ is a descent direction of $\rho$, since according to lemma 13.1:

$$
\rho^{\prime}\left(x_{k} ; \widehat{r}_{k}\right)=(\|\cdot\|)^{\prime}\left(c_{k} ; A_{k} \widehat{r}_{k}\right)=(\|\cdot\|)^{\prime}\left(c_{k} ;-c_{k}\right)=-\left\|c_{k}\right\|<0 .
$$

The second term in the right-hand side of (14.34), $t_{k}:=-Z_{k}^{-} H_{k}^{-1} g_{k}$, is a stationary point of the quadratic problem in $t \in \mathbb{R}^{n}$ :

$$
\left\{\begin{array}{l}
\min _{t} \nabla f_{k}^{\top} t+\frac{1}{2} t^{\top} L_{k} t \\
A_{k} t=0 .
\end{array}\right.
$$

To see this, just set $c_{k}=0$ in (14.8) and (14.34). It is tangent to the manifold $\mathcal{M}_{k}:=\mathcal{M}_{x_{k}}$ and aims at decreasing the function $f$. Indeed, when $H_{k}$ is positive definite and $g_{k} \neq 0, t_{k}$ is a descent direction of $f$ at $x_{k}$, since

$$
f^{\prime}\left(x_{k}\right) \cdot t_{k}=\nabla f\left(x_{k}\right)^{\top}\left(-Z_{k}^{-} H_{k}^{-1} g_{k}\right)=-g_{k}^{\top} H_{k}^{-1} g_{k}<0
$$

We shall come back to this issue in $\S 14.6$, when comparing the direction $d_{k}$ with directions generated by other algorithms.

On the influence of $L_{k}$ on the direction $d_{k}$, we can observe the following.

1. The second-order information used in the Newton direction is entirely contained in the part $Z_{k}^{-\top} L_{k}$ of $L_{k}$. This can be seen in formula (14.34): only this part enters the matrices $\widehat{A}_{k}^{-}$and $H_{k}$. In particular, the direction $d_{k}$ is not changed if we add to $L_{k}$ a matrix of the form $A_{k}^{\top} S_{k} A_{k}$, where $S_{k}$ is an arbitrary symmetric $m \times m$ matrix.
2. If we multiply $L_{k}$ by a number $\alpha \neq 0$, the transversal part $-\widehat{A}_{k}^{-} c_{k}$ of the direction is not affected, while the longitudinal part $-Z_{k}^{-} H_{k}^{-1} g_{k}$ is divided by $\alpha$. In other words, the "size" of $L_{k}$ only acts on the tangential part of $d_{k}$.
Consider now the computation of $\lambda_{k}^{\mathrm{QP}}$. Premultiply the first equation of (14.26) by $A_{k}^{-\top}$ and use formula (14.34) of $d_{k}$ to find

$$
\begin{equation*}
\lambda_{k}^{\mathrm{QP}}=-\widehat{A}_{k}^{-\top} \nabla f_{k}+A_{k}^{-\top} L_{k} \widehat{A}_{k}^{-} c_{k} \tag{14.35}
\end{equation*}
$$

This multiplier, as well as the first term in (14.35),

$$
\begin{equation*}
\widehat{\lambda}_{k}:=-\widehat{A}_{k}^{-\top} \nabla f_{k}, \tag{14.36}
\end{equation*}
$$

are sometimes called second-order multipliers, since they involve second-order derivatives of the functions $f$ and $c$, via the Hessian of the Lagrangian $L_{k}$. These are estimates of the optimal multiplier, since $\lambda_{k}^{\mathrm{QP}}=\widehat{\lambda}_{k}=\lambda_{*}$ when $x_{k}=x_{*}$, a stationary point. Such is also the case of

$$
\lambda_{k}^{\mathrm{LS}}:=-A_{k}^{-\top} \nabla f_{k},
$$

called the first-order multiplier or least-squares multiplier (see (14.24)). It is said to be of first-order because it only involves the first derivatives of the data.

With this section, we have concluded the description of Newton's algorithm to solve equality constrained optimization problems. Next comes the description of an algorithm, also proceeding by linearizations, but different from Newton's method. It can be seen as a kind of nonlinear block GaussSeidel approach.

### 14.5 Reduced Hessian Algorithm

There is an algorithm to solve problem $\left(P_{E}\right)$, different from Newton's method, that also enjoys local quadratic convergence. In optimization, its existence can be suggested by the following considerations.

When $c$ is a submersion on $\Omega$, the feasible set

$$
\mathcal{M}_{*}=\{x \in \Omega: c(x)=0\}
$$

is a manifold of dimension $n-m$. Then $\left(P_{E}\right)$ has only $n-m$ degrees of freedom and a natural question is whether there exists a method where the matrix containing the second-order information (second derivatives of $f$ and $c$ or their quasi-Newton approximation) is only $(n-m) \times(n-m)$. This is certainly the case if the iterates $x_{k}$ are forced to stay in $\mathcal{M}_{*}$. Indeed, such an algorithm can be obtained by taking a parametrization of $\mathcal{M}_{*}$ around $x_{*}$ and applying Newton's method in the parameter space, which has dimension $n-m$. However, requiring $x_{k} \in \mathcal{M}_{*}$ is not realistic: it is often computationally expensive and, anyway, it cannot be realized exactly when $c$ is an arbitrary nonlinear function. What is desired is a method with the following properties:

- the only matrix containing second-order information is $(n-m) \times(n-m)$,
- the iterates $x_{k}$ are not forced to satisfy the constraints at each iteration,
- the speed of convergence is quadratic.

In this section, we show how to introduce such an algorithm. We shall see that this approach is particularly attractive when $n-m \ll n$ and quasiNewton techniques are employed. Throughout the section, we assume that the stationary point $x_{*}$ we are seeking is regular (see definition 14.2).

## The Reduced Optimality System

The first stage leading to the definition of the algorithm is to provide an optimality system of reduced size, with fewer equations than in (14.1). This stage is optional but, by eliminating the multiplier from (14.1), it leads to a more concise presentation.

Premultiply the first equation of (14.1) by $Z^{-}\left(x_{*}\right)^{\top}$ to find, with (14.10), the reduced optimality system:

$$
\left\{\begin{array}{l}
g\left(x_{*}\right)=0  \tag{14.37}\\
c\left(x_{*}\right)=0
\end{array}\right.
$$

where $g$ is the reduced gradient of $f$, defined by (14.31). The multiplier $\lambda_{*}$ no longer appears in this system, which counts $(n-m)+m=n$ equations for the $n$ unknowns $x_{*}$.

Note that the two systems (14.1) and (14.37) have the same solutions $x_{*}$. Indeed, we have just shown that (14.37) can be obtained from (14.1). On the other hand, we deduce from the first equation of (14.37) that

$$
\nabla f\left(x_{*}\right) \in N\left(Z^{-}\left(x_{*}\right)^{\top}\right)=R\left(Z^{-}\left(x_{*}\right)\right)^{\perp}=N\left(A\left(x_{*}\right)\right)^{\perp}=R\left(A\left(x_{*}\right)^{\top}\right)
$$

Therefore there exists $\lambda_{*} \in \mathbb{R}^{m}$ such that $\nabla f\left(x_{*}\right)+A\left(x_{*}\right)^{\top} \lambda_{*}=0$. This is the first equation of (14.1). Thus, there is no loss of solutions by considering (14.37) instead of (14.1).

## Solving the Reduced Optimality System by a Decoupling Technique

The reduced Hessian method essentially consists in performing one Newtonlike step to solve the second equation of (14.37), followed by one Newton-like step to solve the first equation. This resembles a nonlinear block Gauss-Seidel method. There is an important difference however. We shall show that, to yield local quadratic convergence, the first step can be an arbitrary Newtonlike step, but the second one must have a very specific form. In particular, this second step must be tangent to the manifold defined by the second equation in (14.37).

The algorithm generates two sequences of iterates, $\left\{x_{k}\right\}$ and $\left\{y_{k}\right\}$, both converging to the same solution $x_{*}$. Local convergence is studied more easily if the method is thought of generating the sequence $\left\{y_{k}\right\}$. It is this sequence that converges (almost) quadratically. Curiously the sequence $\left\{x_{k}\right\}$ converges slightly less rapidly, but the algorithm is easier to implement in terms of the sequence $\left\{x_{k}\right\}$. We now introduce the method by considering the sequence $\left\{y_{k}\right\}$, while $\left\{x_{k}\right\}$ appears as an intermediate sequence.

Starting with an iterate $y_{k} \in \Omega$, we first perform a Newton-like step that aims at solving the second equation of (14.37). For this, we use a right inverse of the Jacobian of $c$. This gives an intermediate point $x_{k}$, defined by

$$
x_{k}=y_{k}-A^{-}\left(y_{k}\right) c\left(y_{k}\right)
$$

Note that, if $m=n$, then $A^{-}\left(y_{k}\right)$ is the inverse of $A\left(y_{k}\right)$ and the step $-A^{-}\left(y_{k}\right) c\left(y_{k}\right)$ is exactly the Newton step at $y_{k}$ to solve $c(x)=0$ (compare with (13.17) and (13.19)). When $m<n$, which is our situation, every right inverse $A^{-}\left(y_{k}\right)$ produces a particular solution $x_{k}-y_{k}$ to the constraint equation, linearized at $y_{k}$.

We are now interested in making a Newton-like step from $x_{k}$ to solve the first equation of (14.37). The point $x_{k}$ is supposed to be in $\Omega$. Observe first that the reduced gradient can be written $g(x)=Z_{x}^{-\top} \nabla_{x} \ell\left(x, \lambda_{*}\right)$, where $\lambda_{*}$ is the multiplier associated with the solution $x_{*}$. By optimality, $\nabla_{x} \ell\left(x_{*}, \lambda_{*}\right)=0$. Hence, assuming that $Z_{x}^{-}$is continuous at $x_{*}$ and using lemma 13.2, one has

$$
\begin{equation*}
g^{\prime}\left(x_{*}\right)=Z_{*}^{-\top} L_{*}, \tag{14.38}
\end{equation*}
$$

where we have set $Z_{*}^{-}=Z^{-}\left(x_{*}\right)$ and $L_{*}=L\left(x_{*}, \lambda_{*}\right)$, as usual. If ( $x_{*}, \lambda_{*}$ ) is a regular stationary point, the reduced Hessian of the Lagrangian at $\left(x_{*}, \lambda_{*}\right)$,

$$
H_{*}:=Z_{*}^{-\top} L_{*} Z_{*}^{-},
$$

is nonsingular (see proposition 14.1), so that $g^{\prime}\left(x_{*}\right)$ is surjective. Therefore $g$ is a submersion in a neighborhood of $x_{*}$, which is supposed to contain $\Omega$. As above, we can therefore take a right inverse $B^{-}\left(x_{k}\right)$ of $g^{\prime}\left(x_{k}\right)$ and define the next iterate by

$$
y_{k+1}=x_{k}-B^{-}\left(x_{k}\right) g\left(x_{k}\right)
$$

We have just described a procedure for computing $y_{k+1}$ from $y_{k}$, with $x_{k}$ as an intermediate iterate.

We now raise the following question. Is it possible to find a matrix mapping $x \mapsto B^{-}(x)$ so as to obtain fast convergence of the sequence $\left\{y_{k}\right\}$ to $x_{*}$ ? To answer this question, we introduce the functions $\varphi$ and $\psi: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ defined by:

$$
\begin{aligned}
& \varphi(y)=y-A^{-}(y) c(y) \\
& \psi(x)=x-B^{-}(x) g(x) .
\end{aligned}
$$

Then, the procedure we are analyzing can be viewed as fixed point iterations: $y_{k+1}=(\psi \circ \varphi)\left(y_{k}\right)$. As a result, if $B^{-}(\cdot)$ can be determined in such a way that $(\psi \circ \varphi)^{\prime}\left(x_{*}\right)=0$, the algorithm is likely to converge quadratically (see exercise 14.10). The next lemma specifies the value of $B_{*}^{-}:=B^{-}\left(x_{*}\right)$ to get this property.

Lemma 14.6 (condition of quadratic convergence of a decoupling method). Suppose that $g$ and $c$ are differentiable at $x_{*}$, that $A^{-}(\cdot)$ and $B^{-}(\cdot)$ are continuous at $x_{*}$, and that $\left(x_{*}, \lambda_{*}\right)$ is a regular stationary point of $\left(P_{E}\right)$. Then

$$
(\psi \circ \varphi)^{\prime}\left(x_{*}\right)=0 \quad \Longleftrightarrow \quad B_{*}^{-}=Z_{*}^{-} H_{*}^{-1}
$$

where $H_{*}:=Z_{*}^{-\top} L_{*} Z_{*}^{-}$, for some basis $Z_{*}^{-}$of $N\left(A_{*}\right)$.
Proof. Set $B_{*}=Z_{*}^{-\top} L_{*}$ and $C_{*}=(\psi \circ \varphi)^{\prime}\left(x_{*}\right)$. Then, with the assumptions and lemma 13.2:

$$
C_{*}=\left(I-B_{*}^{-} B_{*}\right)\left(I-A_{*}^{-} A_{*}\right)
$$

If $(\psi \circ \varphi)^{\prime}\left(x_{*}\right)=0$, then $C_{*} Z_{*}^{-}=0$, which gives

$$
B_{*}^{-} B_{*} Z_{*}^{-}=Z_{*}^{-} .
$$

We deduce $B_{*}^{-}=Z_{*}^{-} H_{*}^{-1}$. Conversely, if $B_{*}^{-}=Z_{*}^{-} H_{*}^{-1}$, we have $A_{*} B_{*}^{-}=0$. Then

$$
\binom{A_{*}}{B_{*}} C_{*}=0 .
$$

Since the operator applied to $C_{*}$ is nonsingular, we have $C_{*}=(\psi \circ \varphi)^{\prime}\left(x_{*}\right)=0$.

## The Algorithm

Lemma 14.6 suggests designing the algorithm that generates the sequences $\left\{x_{k}\right\}$ and $\left\{y_{k}\right\}$ as follows:

$$
\begin{aligned}
x_{k} & =y_{k}-A^{-}\left(y_{k}\right) c\left(y_{k}\right) \\
y_{k+1} & =x_{k}-Z^{-}\left(x_{k}\right) H_{k}^{-1} g\left(x_{k}\right)
\end{aligned}
$$

Here $H_{k}$ is an $(n-m) \times(n-m)$ matrix approximating the reduced Hessian $H_{*}$ of the Lagrangian, or $Z^{-}\left(x_{k}\right)^{\top} L\left(x_{k}, \lambda_{k}\right) Z^{-}\left(x_{k}\right)$, for a certain multiplier $\lambda_{k}$.

As such, this algorithm can be very time-consuming because the constraints must be linearized at the two points $x_{k}$ and $y_{k}$, and also the two right inverses $A^{-}\left(y_{k}\right)$ and $Z^{-}\left(x_{k}\right)$ must be computed. Even though it is crucial to compute $g$ at $x_{k}$ and $c$ at $y_{k}$, theorem 13.6 states that good convergence can be preserved if the operators involving first derivatives are evaluated at other points; the important thing is that these points converge to the solution. Since the reduced gradient must be evaluated at $x_{k}$, and since it involves a basis $Z^{-}\left(x_{k}\right)$ of the tangent space, the constraints must be linearized at $x_{k}$ anyway. However, $A^{-}$can be evaluated at $x_{k}$ instead of $y_{k}$. This avoids linearizing the constraints at $y_{k}$. Stating the algorithm in terms of the sequence $\left\{x_{k}\right\}$, we then obtain

$$
\begin{aligned}
& y_{k+1}=x_{k}-Z^{-}\left(x_{k}\right) H_{k}^{-1} g\left(x_{k}\right) \\
& x_{k+1}=y_{k+1}-A^{-}\left(x_{k}\right) c\left(y_{k+1}\right) .
\end{aligned}
$$

Finally, setting $g_{k}=g\left(x_{k}\right), A_{k}^{-}=A^{-}\left(x_{k}\right), Z_{k}^{-}=Z^{-}\left(x_{k}\right)$ and

$$
\begin{equation*}
t_{k}=-Z_{k}^{-} H_{k}^{-1} g_{k} \tag{14.39}
\end{equation*}
$$

the algorithm can be stated in a very concise manner:

$$
\begin{equation*}
x_{k+1}=x_{k}+t_{k}-A_{k}^{-} c\left(x_{k}+t_{k}\right) \tag{14.40}
\end{equation*}
$$

As with the Newton method (14.34), the first phase of Algorithm (14.40) consists in performing a displacement tangent to the manifold $\mathcal{M}_{k}$ at $x_{k}$. In the second phase, the algorithm aims at getting the next iterate $x_{k+1}$ closer to the manifold $\mathcal{M}_{*}$ by taking the displacement $-A_{k}^{-} c\left(x_{k}+t_{k}\right)$, in which the constraints are evaluated at $x_{k}+t_{k}$, after the tangent step.

Although the reduced Hessian algorithm, which is summarized in the recurrence (14.40), should be quite clear, we formally state it below.

Reduced Hessian algorithm for $\left(P_{E}\right)$ :
Choose an initial iterate $x_{1}=y_{1} \in \mathbb{R}^{n}$.
Compute $c\left(x_{1}\right), \nabla f\left(x_{1}\right)$, and $A\left(x_{1}\right)$.
Set $k=1$.

1. Compute the reduced gradient $g\left(x_{k}\right)$ by (14.31).
2. Stop if $g\left(x_{k}\right)=0$ and $c\left(y_{k}\right)=0$ (optimality is reached).
3. Compute the reduced Hessian of the Lagrangian $H_{k}$, or an approximation to it, and the tangent step $t_{k}$ by (14.39).
4. Evaluate the constraint at $y_{k+1}:=x_{k}+t_{k}$.
5. Compute the new iterate $x_{k+1}$ by (14.40), $\nabla f\left(x_{k+1}\right)$ and $A\left(x_{k+1}\right)$.

6 . Increase $k$ by 1 and go to 1 .

Note that this algorithm is essentially primal, since it can be expressed only in terms of the primal sequence $\left\{x_{k}\right\}$. A multiplier estimate $\lambda_{k}$ is however often necessary, either to evaluate the reduced Hessian of the Lagrangian at $\left(x_{k}, \lambda_{k}\right)$ in step 3 or to update a quasi-Newton approximation to it (see chapter 18). The cheapest one is the least-squares multiplier defined by (14.24).

## Simplified Newton Method

Algorithm (14.40) would be simpler if, in the second phase, the constraints were evaluated at $x_{k}$. It would then be written

$$
\begin{equation*}
x_{k+1}=x_{k}+t_{k}-A_{k}^{-} c_{k} \tag{14.41}
\end{equation*}
$$

This algorithm is sometimes called the simplified Newton method because it only uses the reduced Hessian of the Lagrangian $H_{k}$, not the full Hessian $L_{k}$ (compare with (14.34) or see $\S 14.6)$. It has a slower convergence speed than (14.40): under natural assumptions, $\left\{x_{k}\right\}$ converges quadratically in two steps (see exercise 14.11). On the other hand, there are examples showing that the sequence $\left\{x_{k}\right\}$ may not converge quadratically in one step (see $[63,373]$ ). To get good convergence, it is therefore important to evaluate the constraints at $x_{k}+t_{k}$, after the tangent displacement.

## Local Convergence

The next theorem states that the sequence $\left\{y_{k}\right\} \equiv\left\{x_{k}+t_{k}\right\}$ of Algorithm (14.40) converges superlinearly if the matrix $H_{k}$ appearing in the tangent step $t_{k}$ satisfies the estimate

$$
H_{k}-H_{*}=O\left(\left\|x_{k}-x_{*}\right\|\right) .
$$

If $H_{k}$ is set to $Z^{-}\left(x_{k}\right)^{\top} L\left(x_{k}, \lambda_{k}\right) Z^{-}\left(x_{k}\right)$, it depends on $x_{k}$ and $\lambda_{k}$ and this condition is satisfied if $\lambda_{k}-\lambda_{*}=O\left(\left\|x_{k}-x_{*}\right\|\right)$ and if the functions $f^{\prime \prime}, c^{\prime \prime}$, and $Z^{-}$are Lipschitzian near $x_{*}$. This leaves a certain freedom for the choice of the multiplier $\lambda_{k}$. For example, one can take $\lambda_{k}=\lambda^{\mathrm{LS}}\left(x_{k}\right)$, the least-squares multiplier defined by (14.24). It is easy to check that $\lambda^{\mathrm{LS}}\left(x_{*}\right)=\lambda_{*}$, and thus $\lambda^{\mathrm{LS}}\left(x_{k}\right)-\lambda_{*}=O\left(\left\|x_{k}-x_{*}\right\|\right)$ if $A^{-}$and $f^{\prime}$ are Lipschitzian near $x_{*}$. With this value of the multiplier, Algorithm (14.40) becomes entirely primal, in the sense that the algorithm only constructs the sequence $\left\{x_{k}\right\}$, the multiplier being reduced to an auxiliary vector, itself depending on $x_{k}$.

The result given below is slightly weaker than theorem 14.5 stating the convergence of $\left\{x_{k}\right\}$ in the primal variant of Newton's algorithm. For that algorithm, the sequence $\left\{x_{k}\right\}$ converges quadratically if $\lambda_{k}-\lambda_{*}=O\left(\left\|x_{k}-x_{*}\right\|\right)$.

The proof of theorem 14.7 uses the notation $O(\cdot)$ as explained at the end of $\S 13.5$. At first, it may disconcert the reader. For example, the first estimate obtained in the proof, namely (14.43), means that there exists a positive constant $C$ such that, if $x_{k}$ is in some neighborhood of $x_{*}$ :

$$
\left\|y_{k+1}-x_{*}-\left(x_{k}-x_{*}\right)+Z_{k}^{-} H_{k}^{-1} Z_{*}^{-\top} L_{*}\left(x_{k}-x_{*}\right)\right\| \leq C\left\|x_{k}-x_{*}\right\|^{2}
$$

The point $x_{k}$ is considered as an arbitrary point in that neighborhood and, despite the presence of the iteration index $k$, there is no reference to a particular sequence. The estimate obtained at the end of the proof, namely $x_{k+2}-x_{*}=O\left(\left\|x_{k}-x_{*}\right\|^{2}\right)$, implies that if $x_{1}$ and $x_{2}$ are in a sufficiently small neighborhood of $x_{*}$, then $x_{3}$ and $x_{4}$ are in that neighborhood (because for example $\left\|x_{3}-x_{*}\right\| \leq\left(C\left\|x_{1}-x_{*}\right\|\right)\left\|x_{1}-x_{*}\right\| \leq\left\|x_{1}-x_{*}\right\|$ if $\left\|x_{1}-x_{*}\right\|$ is sufficiently small). Therefore, by induction, all the estimates can now be applied to all the generated sequences. The interest of this notation is to provide very concise proofs (for another example, see exercise 14.11).

Theorem 14.7 (convergence of the reduced Hessian algorithm). Suppose that $f$ and $c$ are twice differentiable at a regular stationary point $x_{*}$ of problem $\left(P_{E}\right)$ (this allows the use of the operators $Z^{-}(x)$ and $A^{-}(x)$ introduced in §14.2, for $x$ near $x_{*}$ ) and that the reduced gradient $g$ is differentiable near $x_{*}$. Suppose also that $c^{\prime}, g^{\prime}, Z^{-}$and $A^{-}$are Lipschitzian near $x_{*}$, and that the matrix $H_{k}$ used in (14.40) satisfies $H_{k}-H_{*}=O\left(\left\|x_{k}-x_{*}\right\|\right)$. Then, there exists a neighborhood $V$ of $x_{*}$ such that, when the first iterate $x_{1} \in V$, Algorithm (14.40) is well defined and generates a sequence $\left\{x_{k}\right\}$ converging quadratically in two steps to $x_{*}$. Furthermore, the sequence $\left\{y_{k}\right\}$ converges superlinearly to $x_{*}$ with the estimate

$$
\begin{equation*}
y_{k+1}-x_{*}=O\left(\left\|x_{k-1}-x_{*}\right\|\left\|y_{k}-x_{*}\right\|\right) . \tag{14.42}
\end{equation*}
$$

Proof. Remark first that, when $x_{k}$ is close to $x_{*}$, by assumption, $H_{k}$ is close to $H_{*}$, which is nonsingular ( $x_{*}$ is regular). Thus, $H_{k}$ is nonsingular and the iteration is well defined. Also $\left\{H_{k}^{-1}\right\}$ is bounded when $x_{k}$ remains in some neighborhood of $x_{*}$.

Remembering that $y_{k+1}=x_{k}+t_{k}$ and using $g\left(x_{*}\right)=0,(14.38)$, and the Lipschitz continuity of $g^{\prime}$, we have

$$
\begin{align*}
y_{k+1}-x_{*}= & x_{k}-x_{*}-Z_{k}^{-} H_{k}^{-1} g_{k} \\
= & x_{k}-x_{*}-Z_{k}^{-} H_{k}^{-1} Z_{*}^{-\top} L_{*}\left(x_{k}-x_{*}\right)  \tag{14.43}\\
& +O\left(\left\|x_{k}-x_{*}\right\|^{2}\right)
\end{align*}
$$

But $H_{k}^{-1}-H_{*}^{-1}=-H_{k}^{-1}\left(H_{k}-H_{*}\right) H_{*}^{-1}=O\left(\left\|x_{k}-x_{*}\right\|\right)$, so that, with the Lipschitz continuity of $Z^{-}$, the following holds

$$
\begin{equation*}
y_{k+1}-x_{*}=\left(I-Z_{*}^{-} H_{*}^{-1} Z_{*}^{-\top} L_{*}\right)\left(x_{k}-x_{*}\right)+O\left(\left\|x_{k}-x_{*}\right\|^{2}\right) \tag{14.44}
\end{equation*}
$$

This implies in particular that $y_{k+1}-x_{*}=O\left(\left\|x_{k}-x_{*}\right\|\right)$. We also have $x_{k+1}=$ $y_{k+1}-A_{k}^{-} c\left(y_{k+1}\right)$. Therefore, using successively $c\left(x_{*}\right)=0$, the Lipschitz continuity of $c^{\prime}$ and $A^{-},(14.14),(14.44)$, and (14.13), we obtain

$$
\begin{align*}
x_{k+1}-x_{*} & =y_{k+1}-x_{*}-A_{k}^{-} A_{*}\left(y_{k+1}-x_{*}\right)+O\left(\left\|y_{k+1}-x_{*}\right\|^{2}\right) \\
& =y_{k+1}-x_{*}-A_{*}^{-} A_{*}\left(y_{k+1}-x_{*}\right)+O\left(\left\|x_{k}-x_{*}\right\|\left\|y_{k+1}-x_{*}\right\|\right) \\
& =Z_{*}^{-} Z_{*}\left(y_{k+1}-x_{*}\right)+O\left(\left\|x_{k}-x_{*}\right\|\left\|y_{k+1}-x_{*}\right\|\right)  \tag{14.45}\\
& =Z_{*}^{-}\left(Z_{*}-H_{*}^{-1} Z_{*}^{-\top} L_{*}\right)\left(x_{k}-x_{*}\right)+O\left(\left\|x_{k}-x_{*}\right\|^{2}\right) \tag{14.46}
\end{align*}
$$

The operator acting on $\left(x_{k}-x_{*}\right)$ in (14.46) is nonzero in general but its square vanishes, because $\left(Z_{*}-H_{*}^{-1} Z_{*}^{-\top} L_{*}\right) Z_{*}^{-}=0$. From this observation, we deduce the estimate

$$
x_{k+2}-x_{*}=O\left(\left\|x_{k}-x_{*}\right\|^{2}\right)
$$

which shows the two-step quadratic convergence of the sequence $\left\{x_{k}\right\}$.
Using (14.44), (14.45) (at the previous iteration), and observing that

$$
\left(I-Z_{*}^{-} H_{*}^{-1} Z_{*}^{-\top} L_{*}\right) Z_{*}^{-}=0,
$$

we obtain (14.42). The superlinear convergence of $\left\{y_{k}\right\}$ follows.

At this point it is reasonable to wonder why the convergence of the sequence $\left\{y_{k}\right\}$ is not quadratic. Since Algorithm (14.40) uses the second derivatives of $f$ and $c$, it is legitimate to expect quadratic convergence. The above proof clarifies this, indeed: the constraints are not linearized at $y_{k}$, but at the neighboring points $x_{k-1}$ and $x_{k}$. Then, passing from $y_{k}$ to $y_{k+1}$ involves the right inverse $A^{-}\left(x_{k-1}\right)$ instead of $A^{-}\left(y_{k}\right)$, which perturbs the speed of convergence. If the right inverse $A^{-}\left(y_{k+1}\right)$ were used in place of $A^{-}\left(x_{k}\right)$, an $O\left(\left\|y_{k+1}-x_{*}\right\|^{2}\right)$ would appear in (14.45) instead of an $O\left(\left\|x_{k}-x_{*}\right\|\left\|y_{k+1}-x_{*}\right\|\right)$ and quadratic convergence would ensue. Numerically, it is not clear that the computing time of $A\left(y_{k}\right)$ and $A^{-}\left(y_{k}\right)$ would be balanced by the quadratic convergence thus recovered, which is why the algorithm is often stated in the form (14.40).

Beware of the different behavior of the sequences $\left\{x_{k}\right\}$ and $\left\{y_{k}\right\}$. Even though they are generated by the same algorithm and both converge to the same point $x_{*}$, the first one is slower than the second one. This may look surprising, but examples do exist, in which the sequence $\left\{x_{k}\right\}$ does not converge quadratically (see [63]).

## Newton and Quasi-Newton Versions

We have already mentioned that the reduced Hessian method is a very attractive approach when $n-m$ is much smaller than $n$. This is particularly true
for their quasi-Newton versions. In these algorithms the $(n-m) \times(n-m)$ reduced Hessian $H_{k}=Z^{-}\left(x_{k}\right)^{\top} L\left(x_{k}, \lambda_{k}\right) Z^{-}\left(x_{k}\right)$ is approximated by a matrix updated by a quasi-Newton formula (see chapters 4.4 and 18). Only this "small" matrix needs to be updated to collect all the necessary second-order information on the problem that provides superlinear convergence. Furthermore, the small order of these updated matrices makes it possible to rapidly obtain a good approximation of the reduced Hessian.

In the Newton version, $H_{k}$ must be computed. The interest of the reduced Hessian method is then less clear. One way of computing $H_{k}$ is to evaluate first $L\left(x_{k}, \lambda_{k}\right) Z^{-}\left(x_{k}\right)$, by computing $n-m$ directional derivatives of the gradient of the Lagrangian along the columns of $Z^{-}\left(x_{k}\right)$, and then premultiplying the matrix thus obtained by $Z^{-}\left(x_{k}\right)^{\top}$. This computation is conceivable, but the knowledge of $L\left(x_{k}, \lambda_{k}\right) Z^{-}\left(x_{k}\right)$ would allow the use of Newton's method, which does not require any other information on the Hessian of the Lagrangian (see remark 1 on page 235); furthermore, Newton's method does not require a re-evaluation of the constraints after the tangent step.

Another way of getting second-order information in the reduced Hessian algorithm is to approximate $H_{k}$ by computing the directional derivatives of the reduced gradient $g$ along the $n-m$ columns of $Z^{-}\left(x_{k}\right)$. Note that $\tilde{H}_{k}:=g^{\prime}\left(x_{k}\right) Z^{-}\left(x_{k}\right)$ is usually different from $H_{k}$, although, in view of formula (14.38), $g^{\prime}\left(x_{*}\right) Z_{*}^{-}$does equal $Z_{*}^{-\top} L_{*} Z_{*}^{-}$. Now $\tilde{H}_{k}$ satisfies the estimate $\tilde{H}_{k}-H_{*}=O\left(\left\|x_{k}-x_{*}\right\|\right)$ (with sufficiently smooth data), so that theorem 14.7 can be applied. Note also that $\tilde{H}_{k}$ is not necessarily a symmetric matrix. This property depends in particular on the choice of the bases $Z^{-}$: if $Z^{-}(x)$ is computed by partitioning $A(x)$ (i.e., using formula (14.15)), then $\tilde{H}_{k}$ is symmetric; but in general it is not so when orthonormal bases are used (see [149]).

### 14.6 A Comparison of the Algorithms

Table 14.1 and figure 14.4 compare the form and speed of convergence of the three algorithms described in this chapter: Newton (14.6) with (14.5) or (14.34)-(14.35), simplified Newton (14.41), and reduced Hessian (14.40).

In all algorithms, the longitudinal step (tangent to the manifold $\mathcal{M}_{k}$ ) is identical and is written

$$
t_{k}=-Z_{k}^{-} H_{k}^{-1} g_{k}
$$

When $H_{k}$ is positive definite, this step is opposite to the gradient of $f$, seen as a function defined on the manifold $\mathcal{M}_{k}$ equipped at $x_{k}$ with the scalar product (Riemannian structure on $\mathcal{M}_{k}$ ):

$$
\gamma_{x_{k}}\left(Z_{x_{k}}^{-} u, Z_{x_{k}}^{-} v\right)=u^{\top} H_{k} v .
$$

| Algorithms | Longitudinal <br> displacement | Transversal <br> displacement | Speed of <br> convergence |
| :---: | :---: | :---: | :---: |
| Newton | $t_{k}$ | $-\widehat{A}_{k}^{-} c_{k}$ | quadratic |
| Simplified Newton | $t_{k}$ | $-A_{k}^{-} c_{k}$ | 2-step quadratic |
| Reduced Hessian | $t_{k}$ | $-A_{k}^{-} c\left(x_{k}+t_{k}\right)$ | "almost" quadratic |

Table 14.1. Comparison of local methods


Fig. 14.4. Comparison of the Newton $\left(d_{k}^{\mathbb{N}}\right)$, simplified Newton, and reduced Hessian steps

When $H_{k}$ is set to $Z_{k}^{-\top} L_{k} Z_{k}^{-}$and $H_{k}$ is positive definite, $t_{k}$ can also be viewed as the unique solution to the quadratic problem in $t$ :

$$
\left\{\begin{array}{l}
\min _{t} \nabla f_{k}^{\top} t+\frac{1}{2} t^{\top} L_{k} t \\
A_{k} t=0,
\end{array}\right.
$$

This interpretation shows that $t_{k}$ does not depend on the choice of the basis $Z_{k}^{-}$, despite the use of this matrix in the formula above. The algorithms presented in table 14.1 therefore only differ in the choice of the restoration operator, $A_{k}^{-}$or $\widehat{A}_{k}^{-}$, and in the points where the constraints are evaluated, $x_{k}$ or $x_{k}+t_{k}$.

First let us compare the two forms of Newton's method: standard (step given by (14.34)), and simplified (step given by (14.41)). We see that the two displacements have the same form, but the operator acting on $c_{k}=c\left(x_{k}\right)$ is $\widehat{A}_{k}^{-}$in the first case, and $A_{k}^{-}$in the second (both are right inverses of $A_{k}$ ). It has been observed ( $\S 14.2$ ) that $\widehat{A}_{k}^{-}$only depends on the problem's data (see problem 14.19), while $A_{k}^{-}$is the concern of the user of the algorithm. Theorems 14.4 and 14.5 have shown that the choice $\widehat{A}_{k}^{-}$leads to quadratically convergent methods. On the other hand, it is easy to check that the convergence of $\left\{x_{k}\right\}$ with (14.41) is only two-step quadratic when the right inverse $A_{k}^{-}$is arbitrary: one-step quadratic convergence is never guaranteed (see exercise 14.11). Therefore Newton's method is the most effective. Note finally that one can view the simplified Newton method as an algorithm neglecting the part $Z_{k}^{-\top} L_{k} A_{k}^{-}$of $L_{k}$ in the standard Newton method (see
formula (14.33)). Newton's algorithm gains in efficiency from getting more information on the Hessian of the Lagrangian.

As for the reduced Hessian algorithm (14.40), it is very close to the simplified Newton method (14.41). The algorithms differ in the point at which the constraints are evaluated: $x_{k}+t_{k}$ in (14.40) and $x_{k}$ in (14.41). The reduced Hessian method can thus be viewed as a technique to compensate a possible bad choice of right inverse $A_{k}^{-}$by a re-evaluation of the constraints after the tangent step. As shown by theorem 14.7, this yields a good speed of convergence for the sequence $\left\{x_{k}+t_{k}\right\}$, a property that is not shared with the simplified Newton algorithm.

### 14.7 The Hanging Chain Project II

The goal of the second session is to implement one of the local algorithms introduced in this chapter and to understand its behavior on the hanging chain test problem presented in $\S 13.8$ (we assume here that the main program and the simulator have been written in Matlab). Various algorithms can be implemented. Below, we concentrate our comments on the standard Newton method described on page 221 in $\S 14.1$, because it is this algorithm that is the easiest to extend to inequality constrained problems. We shall gain experience on its features, its efficiency, and shall reveal its weak points (some of them will be fixed in the next chapters).

We refer the reader to figure 13.3 for the general flowchart of the program. In this session, we start to write the optimization function sqp, which is assumed to be in the file sqp.m. We want to have an implementation that can be used to solve other optimization problems than the hanging chain test problem. This is a good reason for using the mathematical notation of this chapter inside sqp.m, not the language linked to the test problem. In our implementation, the function sqp has the following form

```
function [x,lme,lmi,info] = ...
    sqp (simul,x,lme,lmi,f,ce,ci,g,ae,ai,hl,options)
```

Some of the input or output arguments can be empty, depending on the presence of equality and/or inequality constraints; in particular, the variables in connection with the inequality constraints can be ignored for the while. The input arguments are the following: simul is a string giving the name of the simulator (here 'chs'); x is the initial value of the primal variable $x$ (position of the joints); Ime and lmi are the initial values of the multiplier $\lambda_{E}$ and $\lambda_{I}$ associated with the equality and inequality constraints; $f$, ce, and ci are the values of the objective function $f$ to minimize (the energy) and of the equality and inequality constraint functions $c_{E}$ and $c_{I}$ (lengths of the bars and floor constraint) at the initial point $x ; \mathrm{g}, \mathrm{ae}$, and ai are the values of the gradient of $f$ and the Jacobian matrices $A_{E}$ and $A_{I}$ of $c_{E}$ and $c_{I}$ at the
initial point $x$; hl is the Hessian of the Lagrangian at the initial $(x, \lambda)$ or an approximation to it; and the structure options is aimed at tuning the behavior of the solver. Standard options include upper bounds on the number of iterations and simulations (options.iter and options.simul), the required tolerances on the KKT conditions (options.tol(1:4), see below), the output channel for printing (options.fout), etc. Other options will be discussed in other sessions. The output arguments are as follows: x, lme, and lmi are the final values of the primal and dual (multipliers) variables found by sqp; and info is a structure providing various information on the course of the optimization realized by the solver, telling in particular whether optimality has been reached, up to the required precision specified by the options.tol input argument, and in any case the reason why the solver has stopped.

We have already said on page 228 that the Newton algorithm aims at finding a stationary point, i.e., a pair $\left(x_{*}, \lambda_{*}\right)$ satisfying the optimality conditions (13.1), not necessarily a local minimum. Therefore, it makes sense to have a stopping criterion based on these conditions. In our code, we stop the iterations as soon as, for some norms, the current iterate $(x, \lambda)$ satisfies

$$
\begin{aligned}
\left\|\nabla f(x)+A(x)^{\top} \lambda\right\| & \leq \text { options.tol (1) } \\
\left\|c_{E}(x)\right\| & \leq \text { options.tol (2) } \\
\left\|c_{I}(x)^{+}\right\| & \leq \text {options.tol (3) } \\
\max \left(\left\|\lambda_{I}^{-}\right\|,\left\|\Lambda_{I}^{\top} c_{I}(x)\right\|\right) & \leq \text { options.tol (4) }
\end{aligned}
$$

where $t^{+}=\max (0, t), t^{-}=\max (0,-t)$, and $\Lambda_{I}=\operatorname{Diag}\left(\lambda_{I}\right)$.
Writing the Matlab function sqp implementing the Newton algorithm of page 221 is actually extremely simple. The core of the function is only a few lines long. The time consuming operation is the one to solve the linear system in step 2, but for a small problem this is straightforward. The easiest way of doing this operation is to form the matrix $K$ in (14.9) and to use the standard linear solver of Matlab (see § 14.4 for other possibilities). Since hl and ae are the variables containing respectively the Hessian of the Lagrangian and the Jacobian of the equality constraints, steps 2 and 3 of the algorithm are simply made up of

```
K = [hl ae'; ae zeros(me)];
d = -K\ [g;ce];
x = x + d(1:n);
lme = d(n+1:n+me);
```

where me $=m_{E}$ is the number of equality constraints, $\mathrm{n}=n$ is the number of variables, and the final values of x and lme are the updated iterates $x_{+}$ and $\lambda_{+}$.

## Algorithmic Details, Errors to Avoid, Difficulties to Overcome

The solver sqp offers the user the possibility to set the initial value of $x$ and $\lambda$. This is interesting when it is desirable to restart the solver from a
known approximate solution (recall that the method is primal-dual so that both $x$ and $\lambda$ must be specified). More generally, requiring to initialize $x$ is sensible, since the user often knows an approximate solution to the problem. This is less clear for $\lambda$, since the multipliers have sometimes a less direct "physical" meaning or, perhaps, this meaning is known but the value of $\lambda$ is still difficult to determine. Therefore, it is sometimes wise to let the solver choose the initial multiplier. For an equality constrained problem, one often computes the initial $\lambda$ as the solution to the linear least-squares problem

$$
\begin{equation*}
\min _{\lambda \in \mathbb{R}^{m}} \frac{1}{2}\left\|\nabla_{x} \ell(x, \lambda)\right\|_{2}^{2} \tag{14.47}
\end{equation*}
$$

This is motivated by the fact that the gradient of the Lagrangian vanishes at a solution. The convex quadratic problem above always has a solution (theorem 19.1), which is the least-squares multiplier (14.24) when $c_{E}^{\prime}(x)$ is surjective.

The Newton algorithm is structured as an iteration loop, which contains the piece of code given above. Of course the simulator simul must be called at each iteration after having computed $x_{+}$and $\lambda_{+}$, in order to update the values of $\mathrm{hl}, \mathrm{ae}, \mathrm{g}$, and ce and to check optimality.

Writing an optimization software is a special computer science activity in the sense that the realized code has to control the convergence of a sequence. In some cases, the sequence may diverge simply because the conditions of convergence are not satisfied, not because of an error in the code. Since convergence requires an unpredictable number of iterations, it is sometimes difficult to tell on a particular case whether the behavior of the solver is correct. To certify the correctness of the function sqp, a good idea is to try it on problems with an increasing difficulty and to check the quadratic convergence of the generated sequences, as explained below.

- Try first to start sqp at the solution to a trivial problem: for example, the chain with 2 bars of length 5 , with $(a, b)=(6,0)$, whose single joint should be at position $(3,-4)$. The solver should stop without making any iteration, so that this test case checks only the validity of the stopping criterion and the simulator.
- Try next to start sqp near the solution to an easy problem: for example, the chain with 3 bars of length 5 , with $(a, b)=(11,0)$, whose joints should be at position $(3,-4)$ and $(8,-4)$. Convergence should be obtained in very few iterations, if the initial nodes are at positions $(2,-5)$ and $(9,-3)$. Our code converges in 5 iterations with options.tol (1:4) set to $1 . \mathrm{e}-10$.
The Newton algorithm of page 221 is known to converge quadratically if the initial primal-dual iterate $\left(x_{1}, \lambda_{1}\right)$ is sufficiently close to a regular stationary point (theorem 14.4). Checking that quadratic convergence actually occurs is a good way of verifying that the implementation of both the algorithm and the simulator has been done properly. The very definition of
quadratic convergence of a sequence $\left\{z_{k}\right\}$ makes use of the limit point $z_{*}$ to which it converges (see $\S 13.5$ ). Since, in the course of the optimization, the limit point $z_{*}:=\left(x_{*}, \lambda_{*}\right)$ of the generated sequence $\left\{z_{k}\right\}:=\left\{\left(x_{k}, \lambda_{k}\right)\right\}$ is not known, the definition cannot be directly applied. The idea is then to observe the behavior of another sequence, whose limit point is zero (hence known!) and that also converges quadratically. Below, we consider the following two possibilities.
- For Newton's method, a natural object to look at is the function of which the algorithm tries to find a zero. For an equality constrained optimization problem, it is the function $z:=(x, \lambda) \in \mathbb{R}^{n+m} \mapsto F(z)=$ $\left(\nabla_{x} \ell(x, \lambda), c(x)\right) \in \mathbb{R}^{n+m}$. When $z_{*}:=\left(x_{*}, \lambda_{*}\right)$ is a regular stationary point (definition 14.2), $F^{\prime}\left(z_{*}\right)$ is nonsingular and it is not difficult to show that $\left(z_{k}-z_{*}\right) \sim F\left(z_{k}\right)$ in the sense of (13.11). Therefore $F\left(z_{k}\right) \rightarrow 0$ quadratically in Newton's algorithm.
- Another vector that tends to zero is the step $s_{k}:=z_{k+1}-z_{k}$. By lemma 13.5, $\left\{s_{k}\right\}$ also converges quadratically to zero in Newton's method.

Let us check quadratic convergence of our implementation on the following test case.

Test case 1a: second hook at $(a, b)=(1,-0.3)$, lengths of the bars: $L=$ $(0.4,0.3,0.25,0.2,0.4)$, and initial positions of the chain joints: $(0.2,-0.5)$, $(0.4,-0.6),(0.6,-0.8)$, and $(0.8,-0.6)$.
The results obtained with test case 1a are shown in figure 14.5. Convergence


Fig. 14.5. Test case 1 a
with options.tol $(1: 4)=10^{-10}$ is obtained in 6 iterations. The picture on the left shows the initial position of the chain (thin solid bars), the 5 intermediate positions (dashed bars) and the final position (bold solid bars). The picture on the right gives a plot of the ratios $\left\|F\left(z_{k+1}\right)\right\|_{2} /\left\|F\left(z_{k}\right)\right\|_{2}^{2}$ and $\left\|s_{k+1}\right\|_{2} /\left\|s_{k}\right\|_{2}^{2}$, for $k=1, \ldots, 5$. The boundedness of these ratios leaves no doubt on the quadratic convergence of the sequence $\left\{z_{k}\right\}$ to its limit.

## Experimenting with the Newton Method

The test case 1a reveals the ideal behavior of Newton's method: quadratic convergence is obtained when the initial position of the chain is close to a regular solution. This solution is a strict local minimum (the smallest eigenvalue of the reduced Hessian of the Lagrangian $Z_{*}^{-\top} L_{*} Z_{*}^{-}$, for some orthonormal basis $Z_{*}^{-}$, is positive) and probably the global one.

Other solutions can be found by Newton's method with the same data, and those are not local minima. This is the case with the following two starting points.

Test case 1b: identical to test case 1a, except that the initial positions of the chain joints are $(0.2,0.5),(0.4,0.6),(0.6,0.8)$, and $(0.8,0.6)$.

Test case 1c: identical to test case 1a, except that the second hook at $(a, b)=$ $(0.8,-0.3)$ and that the initial positions of the chain joints are $(0.3,0.3)$, $(0.5,0.4),(0.3,0.4)$, and $(0.6,0.3)$.
The resulting equilibria are shown in figure 14.6. The picture on the left


Fig. 14.6. Test cases 1b and 1c: a maximum (left) and a stationary point (right)
shows a local maximum (the largest eigenvalue of the reduced Hessian of the Lagrangian is negative). The right hand side picture shows a stationary point that is neither a minimum nor a maximum (the $3 \times 3$ reduced Hessian of the Lagrangian has two negative eigenvalues and a positive one).

The next two examples have been built to show cases without convergence.
Test case 1d: identical to test case 1a, except that the initial positions of the chain joints are $(0.2,-0.5),(0.4,1.0),(0.6,-0.8)$, and $(0.8,-0.6)$ (hence, only the $y$-coordinate of the second joint has been modified).
Test case 2a: second hook at $(a, b)=(2,0)$, lengths of the bars: $L=(1,1)$, and initial position of the chain joint: $(1.5,-0.5)$.

The results are shown in figure 14.7. In the left picture, we have only plotted


Fig. 14.7. Test cases 1 d , 2 a , and 2 b : non convergence in $(x, \lambda)$ (left), non convergence in $\lambda$ (middle), and convergence in $(x, \lambda)$ (right)
the position of the chain at the first 10 iterations, since apparently Newton's method does not converge. The generated sequence has a typical erratic behavior. By chance, one of these iterates may fall into the neighborhood of convergence of a stationary point, but this does not occur during the first 50 iterations. The middle picture is more puzzling, since it looks as if the algorithm converges. This is actually the case for the primal variables $x$ (giving the position of the chain), which converge to the single feasible joint $(1,0)$, but the dual variables diverge (their norm blows up). This reflects the fact that the optimal solution does not satisfy the KKT conditions (the Jacobian of the equality constraint in not surjective at the solution and there is no optimal multipliers); in fact, a weighty chain formed of two horizontal bars is not physically possible. The situation is quite different for the similar test case 2b below.

Test case 2b: second hook at $(a, b)=(0,-2)$, lengths of the bars: $L=(1,1)$, and initial position of the chain joint: $(0.5,-0.5)$.

The result is shown in the right hand side picture in figure 14.7: convergence in both $(x, \lambda)$ is obtained in 17 iterations.

We conclude with the following test case and let the reader guess whether the position of the chain given in figure 14.8 is a local minimum.

Test case 3: second hook at $(a, b)=(0,-1)$ and lengths of the bars: $L=(0.5$, $0.5,2.0,0.4,0.4)$.

## Notes

The operators $A^{-}, Z^{-}$, and $Z$ defined in $\S 14.2$ were introduced by Gabay [137]. They have allowed us to use the same formalism for the optimal control and orthogonal settings. We have seen that convergence results need to have a smooth map $x \mapsto\left(A_{x}^{-}, Z_{x}^{-}, Z_{x}\right)$. It is usually difficult to guarantee this smoothness in a large region (for example there is no continuous basis


Fig. 14.8. Test case 3: is this a stable static equilibrium position?
mapping $x \mapsto Z_{x}^{-}$on a sphere of even dimension). Even locally, standard procedures such as the QR factorization presented in $\S 14.2$ may compute a noncontinuous basis mapping [83]. This issue has been examined by several authors, who have proposed procedures for computing a smoothly varying sequence of matrices $Z_{k}^{-}$when approaching a solution: see [83, 157, 24, 68]. The connection between the symmetry of $g^{\prime}(x) Z^{-}(x)$ and the choice of basis of the tangent space is discussed in [149; §3].

The accuracy of the computation of the Newton step by the reduced system approach (see §14.4) crucially depends on the choice of operators $A^{-}$ and $Z^{-}$. When these are obtained from the partitioning of $A$ into $(B N)$, with a nonsingular $B$, and from a Gaussian factorization of $B$, Fletcher and Jonhson [129] recommend to use Gaussian elimination on the whole matrix $A^{\top}$ to get

$$
A^{\top}=\binom{L_{1}}{L_{2}} U
$$

where $L_{1}$ is unit lower triangular and $U$ is upper triangular. The elements of $L_{1}$ and $L_{2}$ can be guaranteed to be not bigger than 1 in absolute value (e.g., because the elements of $N^{\top}$ are taken into account in the choice of the pivots). This approach provides well conditioned basis $Z^{-}$and a solution to the Newton system that is less sensitive to the ill-conditioning of $A$ and that of the reduced Hessian of the Lagrangian.

The presentation of the reduced Hessian method given in $\S 14.5$ follows [145]. This algorithm, condensed in formula (14.40), was introduced by Coleman and Conn [81], who proved convergence of the sequence $\left\{x_{k}\right\}$. Superlinear (or quadratic) convergence of the sequence $\left\{y_{k}\right\}$ was observed independently by Hoyer [197], Gilbert [145], and Byrd [64]. The simplified Newton method (14.41) has been studied by many authors: Murray and Wright [271], Powell [292], Gabay [138], Nocedal and Overton [276], Byrd and Nocedal [67], to mention a few. Newton's method on the reduced system (14.37) is considered by Goodman [177], who analyses its links with Newton's algorithm (14.5)-(14.6).

## Exercises

14.1. Nonconvergence with a step computed by (14.2). Consider the problem in $x=\left(x_{1}, x_{2}\right) \in \mathbb{R}^{2}:$

$$
\left\{\begin{array}{l}
\min _{x}-a x_{1}^{2}+2 x_{2} \\
x_{1}^{2}+x_{2}^{2}=1,
\end{array}\right.
$$

where $a \in] 0,1\left[\right.$. Show that the unique solution $x_{*}=(0,-1)$ to this problem can be repulsive for an algorithm based on (14.2): for $x$ on the constraint manifold, arbitrary close to (but different from) the solution, and for a stationary point $d$ of (14.2), $x+d$ is further from the solution than $x$.
14.2. Inertia of the matrix $K$ in (14.9). The inertia $i$ of a matrix is the triple $\left(n_{-}, n_{0}, n_{+}\right)$formed by the numbers of its negative, null, and positive eigenvalues respectively. Let $K$ be the matrix defined in (14.9), where $L$ is an $n \times n$ symmetric matrix and $A$ is an $m \times n$ surjective matrix (hence $m \leq n$ ). Show that

$$
i(K)=i\left(Z^{-\top} L Z^{-}\right)+(m, 0, m)
$$

where the columns of $Z^{-}$form a basis of $N(A)$ (see [90, 72, 179, 244] for related results).
[Hint: Prove the following claims and conclude: $(i) n_{0}(K)=n_{0}\left(Z^{-\top} L Z^{-}\right)$; (ii) there is no restriction in assuming that $Z^{-\top} L Z^{-}$is nonsingular (use a perturbation argument, for instance), which is supposed from now on; (iii) $i(K)=i\left(Z^{-\top} L Z^{-}\right)+i(\Sigma)$, where

$$
\Sigma:=\left(\begin{array}{cc}
S & I_{m} \\
I_{m} & 0
\end{array}\right)
$$

for some $m \times m$ symmetric matrix $S$ (use the matrix $\widehat{A}^{-}$defined by (14.21) and Sylvester's law of inertia: $i\left(P K P^{\top}\right)=i(K)$ if $P$ is nonsingular); (iv) $i(\Sigma)=(m, 0, m)$.]
14.3. Regular stationary points are isolated. Let $\left(x_{*}, \lambda_{*}\right)$ be a regular stationary point of problem $\left(P_{E}\right)$. Show that there is a neighborhood of $\left(x_{*}, \lambda_{*}\right)$ in $\mathbb{R}^{n} \times \mathbb{R}^{m}$ containing no other stationary point than $\left(x_{*}, \lambda_{*}\right)$.
14.4. A view of the reduced Hessian of the Lagrangian. Let $f: \Omega \rightarrow \mathbb{R}$ and $c: \Omega \rightarrow \mathbb{R}^{m}$ be twice differentiable functions defined in a neighborhood $\Omega$ of a point $x_{*} \in \mathbb{R}^{n}$ and denote $\ell(x, \lambda):=f(x)+\lambda^{\top} c(x)$, for $(x, \lambda) \in \Omega \times \mathbb{R}^{m}$, and $L_{*}:=\nabla_{x x}^{2} \ell\left(x_{*}, \lambda_{*}\right)$. Suppose that $\nabla_{x} \ell\left(x_{*}, \lambda_{*}\right)=0$ for some $\lambda_{*} \in \mathbb{R}^{m}$ (it is not assumed that $\left.c\left(x_{*}\right)=0\right)$ and that $A_{*}:=c^{\prime}\left(x_{*}\right)$ is surjective. Let $Z_{*}^{-}$ be an $n \times(n-m)$ matrix whose columns form a basis of $N\left(A_{*}\right)$. Show that one can find a twice differentiable parametric representation $\varphi: U \subset \mathbb{R}^{n-m}$ $\rightarrow \mathcal{M}_{x_{*}} \subset \mathbb{R}^{n}$ of the manifold $\mathcal{M}_{x_{*}}:=\left\{x \in \Omega: c(x)=c\left(x_{*}\right)\right\}$ around $x_{*}$ defined in a neighborhood $U$ of 0 , such that $\varphi(0)=x_{*}, \nabla(f \circ \varphi)(0)=0$, and $\nabla^{2}(f \circ \varphi)(0)=Z_{*}^{-\top} L_{*} Z_{*}^{-}$is the reduced Hessian of the Lagrangian.
14.5. Right inverse and complementary subspace. Let $A$ be an $m \times n$ surjective matrix and $\mathcal{S}$ be a subspace of $\mathbb{R}^{n}$, complementary to $N(A)$ (i.e., $N(A) \cap \mathcal{S}=$ $\{0\}$ and $\operatorname{dim} \mathcal{S}=m)$. Show that there exists a unique right inverse $A^{-}$of $A$ such that $R\left(A^{-}\right)=\mathcal{S}$.
14.6. On the orthogonal decomposition. Let $A$ be an $m \times n$ surjective matrix, $A^{-}$ be a right inverse of $A$ and $Z^{-}$be a matrix whose columns form a basis of $N(A)$. Show that $A^{-} A+Z^{-} Z^{-\top}=I_{n}$ if and only if $A^{-}=A^{\top}\left(A A^{\top}\right)^{-1}$ (i.e., $A^{-}$is the unique right inverse of $A$ whose range space is perpendicular to $N(A))$ and $Z^{-\top} Z^{-}=I_{n-m}$ (i.e., the columns of $Z^{-}$are orthonormal).
14.7. On the oblique right inverse. Let $A$ be an $m \times n$ surjective matrix. Find an $n \times n$ symmetric matrix $M$, that is positive definite in the null space of $A$, such that the right inverse $\widehat{A}^{-}$of $A$ defined by (14.20) is the one given by formula (14.16). The same question to recover the right inverse given by formula (14.17).
14.8. Quadratic convergence of $\left\{\left(x_{k}, y_{k}\right)\right\}$ without linear convergence of $\left\{x_{k}\right\}$. Let $\left.y_{1} \in\right] 0,1\left[\right.$ and consider the sequence $\left\{\left(x_{k}, y_{k}\right)\right\}_{k \geq 1} \in \mathbb{R}^{2}$ generated by $y_{k+1}=y_{k}^{2}, x_{k+1}=x_{k}$ if $k$ is odd and $x_{k+1}=y_{k+1}^{2}$ if $k$ is even. Show that $\left\{\left(x_{k}, y_{k}\right)\right\}$ converges quadratically to $(0,0)$, while $\left\{x_{k}\right\}$ does not even converge linearly to 0 .
14.9. Least-squares multiplier. Suppose that $A(x)=c^{\prime}(x)$ is surjective and let $A^{-}(x)$ be a right inverse of $A(x)$. Find a least-squares problem, to which the least-squares multiplier $\lambda^{\mathrm{LS}}(x)=-A^{-}(x)^{\top} \nabla f(x)$ is the solution.
[Hint: The least-squares problem has the form $\min _{\lambda \in \mathbb{R}^{m}}\left\|M \nabla_{x} \ell(x, \lambda)\right\|_{2}$, for some nonsingular matrix $M$ to be found.]
14.10. Quadratically convergent fixed point iterations. Let $\Psi: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ be a $C^{1,1}$ map in the neighborhood of one of its fixed points $x_{*}$ (i.e., $\Psi\left(x_{*}\right)=x_{*}$ ). Suppose that $\Psi^{\prime}\left(x_{*}\right)=0$. Show that if $x_{1}$ is sufficiently close to $x_{*}$, then the sequence generated by $x_{k+1}=\Psi\left(x_{k}\right)$, for $k \geq 1$, converges quadratically to $x_{*}$.
14.11. Convergence of the simplified Newton method. Suppose that $f$ and $c$ are twice differentiable at a regular stationary point $x_{*}$ of problem $\left(P_{E}\right)$ (this allows the use of the operators $Z^{-}(x)$ and $A^{-}(x)$ introduced in $\S 14.2$, for $x$ near $x_{*}$ ) and that the reduced gradient $g$ is differentiable near $x_{*}$. Suppose also that $c^{\prime}, g^{\prime}, Z^{-}$and $A^{-}$are Lipschitzian near $x_{*}$, and that the matrix $H_{k}$ used in the simplified Newton method (14.41) satisfies $H_{k}-H_{*}=$ $O\left(\left\|x_{k}-x_{*}\right\|\right)$. Then, there exists a neighborhood $V$ of $x_{*}$ such that, when the first iterate $x_{1} \in V$, Algorithm (14.41) is well defined and generates a sequence $\left\{x_{k}\right\}$ converging quadratically in two steps to $x_{*}$.
[Hint: Show that $x_{k+1}-x_{*}=Z_{*}^{-}\left(Z_{*}-H_{*}^{-1} Z_{*}^{-\top} L_{*}\right)\left(x_{k}-x_{*}\right)+O\left(\left\|x_{k}-x_{*}\right\|^{2}\right)$, applying a technique similar to the one used in the proof of theorem 14.7, and conclude.]

## 15 Local Methods for Problems with Equality and Inequality Constraints

In this chapter, we consider the general minimization problem $\left(P_{E I}\right)$, with equality and inequality nonlinear constraints, which we recall in figure 15.1. The notation used to describe this problem was given in the introduction,

$$
\left(P_{E I}\right)\left\{\begin{array}{l}
\min _{x} f(x) \\
c_{E}(x)=0 \\
c_{I}(x) \leq 0 \\
x \in \Omega
\end{array}\right.
$$



Fig. 15.1. Problem $\left(P_{E I}\right)$ and its feasible set
on page 193. As in chapter 14 , we always suppose that $c_{E}$ is a submersion (i.e., $c_{E}^{\prime}(x)$ is surjective or onto for all $x$ in the open set $\Omega$ ); hence the set $c_{E}^{-1}(0):=\left\{x \in \Omega: c_{E}(x)=0\right\}$ is a submanifold of $\mathbb{R}^{n}$. The feasible set of $\left(P_{E I}\right)$, denoted by

$$
X:=\left\{x \in \Omega: c_{E}(x)=0, c_{I}(x) \leq 0\right\}
$$

is then the part of this manifold formed of the points also satisfying the inequality constraints $c_{i}(x) \leq 0$ for all $i \in I$. The set delimited by the curves of $c_{E}^{-1}(0)$ in figure 15.1 is a typical example of feasible set for problem $\left(P_{E I}\right)$. We have put the solution $x_{*}$ on the boundary of this set, but nothing imposes that this actually occurs. The solution could just as well be inside the curved triangle without touching the solid lines. Finding a solution like the one in figure 15.1 is usually more difficult than when there is no active inequality constraints (and when this fact is known). An additional fearsome difficulty, not present in problem $\left(P_{E}\right)$, is indeed linked to the determination of the active constraints at the solution.

Let us recall the first-order optimality conditions of problem $\left(P_{E I}\right)$ : when the constraints are qualified at a solution $x_{*} \in X$, there exists a Lagrange
multiplier vector $\lambda_{*} \in \mathbb{R}^{m}$ such that

$$
(\mathrm{KKT}) \quad\left\{\begin{array}{l}
(a) \nabla f\left(x_{*}\right)+A\left(x_{*}\right)^{\top} \lambda_{*}=0  \tag{15.1}\\
(b) c_{E}\left(x_{*}\right)=0, \quad c_{I}\left(x_{*}\right) \leq 0 \\
(c)\left(\lambda_{*}\right)_{I} \geq 0 \\
(d)\left(\lambda_{*}\right)_{I}^{\top} c_{I}\left(x_{*}\right)=0
\end{array}\right.
$$

This chapter is organized as follows. In $\S 15.1$, the SQP algorithm is introduced as a Newton-like approach to solve the KKT system (15.1). We shall stress the fact that, in the presence of nonconvexity, the solution to the osculating quadratic problem has to be selected with care. In § 15.2, we give conditions ensuring primal-dual quadratic convergence. First, the case when strict complementarity holds is examined. The active constraints at the solution are shown to be identified by the osculating quadratic problem as soon as the primal-dual iterate is in some neighborhood of a regular stationary point. The algorithm then reduces to Newton's method for the problem where the active constraints are considered as equality constraints, so that the local convergence result of theorem 14.4 can be applied. Next, we focus on the case without strict complementarity and show that quadratic convergence still holds, although the active constraint are no longer necessarily correctly identified by the osculating quadratic program. Necessary and sufficient conditions for primal superlinear convergence are given in $\S 15.3$.

### 15.1 The SQP Algorithm

## Introduction of the Algorithm

The Sequential Quadratic Programming (SQP) algorithm is a form of Newton's method to solve problem $\left(P_{E I}\right)$ that is well adapted to computation. We have seen in chapter 14 that, to introduce such an algorithm, it is a good idea to start with the linearization of the optimality conditions and we follow the same approach here. Let us linearize (15.1) at the current point $\left(x_{k}, \lambda_{k}\right)$, denoting by $\left(d_{k}, \mu_{k}\right)$ the change in the variables. This one solves the following system of equalities and inequalities in the unknown $(d, \mu)$ :

$$
\left\{\begin{array}{l}
L_{k} d+A_{k}^{\top} \mu=-\nabla_{x} \ell_{k}  \tag{15.2}\\
\left(c_{k}+A_{k} d\right)^{\#}=0 \\
\left(\lambda_{k}+\mu\right)_{I} \geq 0 \\
\left(\lambda_{k}+\mu\right)_{I}^{\top}\left(c_{k}\right)_{I}+\left(\lambda_{k}\right)_{I}^{\top}\left(A_{k} d\right)_{I}=0
\end{array}\right.
$$

As before, we use the notation $c_{k}:=c\left(x_{k}\right), A_{k}:=A\left(x_{k}\right):=c^{\prime}\left(x_{k}\right)$, $\nabla_{x} \ell_{k}=\nabla_{x} \ell\left(x_{k}, \lambda_{k}\right)$ and $L_{k}:=\nabla_{x x}^{2} \ell\left(x_{k}, \lambda_{k}\right)$. The notation $(\cdot)^{\#}$ was defined on page 194.

Because of its inequalities, (15.2) is not simple to solve. The key observation is that a good interpretation can be obtained if we add to the last equation the term $(\mu)_{I}^{\top}\left(A_{k} d\right)_{I}$. Compared with the others, this term is negligible
when the steps $\mu_{k}$ and $d_{k}$ are small, which should be the case when the iterates are close to a solution to $\left(P_{E I}\right)$. Introducing the unknown $\lambda^{\mathrm{QP}^{P}}:=\lambda_{k}+\mu$, the modified system (15.2) can then be written

$$
\left\{\begin{array}{l}
L_{k} d+A_{k}^{\top} \lambda^{\mathrm{QP}}=-\nabla f_{k}  \tag{15.3}\\
\left(c_{k}+A_{k} d\right)^{\#}=0 \\
\left(\lambda^{\mathrm{QP}}\right)_{I} \geq 0 \\
\left(\lambda^{\mathrm{QP}}\right)_{I}^{\top}\left(c_{k}+A_{k} d\right)_{I}=0 .
\end{array}\right.
$$

A remarkable fact, easy to check, is that (15.3) is the optimality system of the following osculating quadratic problem (QP)

$$
\left\{\begin{array}{l}
\min _{d} \nabla f\left(x_{k}\right)^{\top} d+\frac{1}{2} d^{\top} L_{k} d  \tag{15.4}\\
c_{E}\left(x_{k}\right)+A_{E}\left(x_{k}\right) d=0 \\
c_{I}\left(x_{k}\right)+A_{I}\left(x_{k}\right) d \leq 0
\end{array}\right.
$$

This QP is easily obtained from $\left(P_{E I}\right)$. Its constraints are those of $\left(P_{E I}\right)$, linearized at $x_{k}$. Its objective function is hybrid, with $\nabla f\left(x_{k}\right)$ in the linear part and the Hessian of the Lagrangian in its quadratic part. The osculating quadratic problem (14.8), associated with the equality constrained problem $\left(P_{E}\right)$, has made us familiar with the structure of (15.4).

We call Sequential Quadratic Programming (SQP) the algorithm generating a sequence $\left\{\left(x_{k}, \lambda_{k}\right)\right\}$ of approximations of $\left(x_{*}, \lambda_{*}\right)$ by computing at each iteration a primal-dual stationary point $\left(d_{k}, \lambda_{k}^{\mathrm{QP}}\right)$ of the quadratic problem (15.4), and by setting $x_{k+1}=x_{k}+d_{k}$ and $\lambda_{k+1}:=\lambda_{k}^{\mathrm{QP}}$.

Sequential Quadratic Programming (SQP):
An initial iterate $\left(x_{1}, \lambda_{1}\right)$ is given.
Compute $c\left(x_{1}\right), \nabla f\left(x_{1}\right)$, and $A\left(x_{1}\right)$.
Set $k=1$.

1. Stop if the KKT conditions (15.1) holds at $\left(x_{*}, \lambda_{*}\right) \equiv\left(x_{k}, \lambda_{k}\right)$ (optimality is reached).
2. Compute $L\left(x_{k}, \lambda_{k}\right)$ and find a primal-dual stationary point of (15.4), i.e., a solution $\left(d_{k}, \lambda_{k}^{\mathrm{QP}}\right)$ to (15.3).
3. Set $x_{k+1}:=x_{k}+d_{k}$ and $\lambda_{k+1}:=\lambda_{k}^{\mathrm{QP}}$.
4. Compute $c\left(x_{k+1}\right), \nabla f\left(x_{k+1}\right)$, and $A\left(x_{k+1}\right)$.

5 . Increase $k$ by 1 and go to 1 .

This algorithm assumes that the QP (15.4) always has a solution or, equivalently, that it is feasible and bounded (theorem 19.1). Adapted remedies must be implemented when this does not happen, such as the elastic mode of [156], which deals with infeasible linearized constraints.

What is gained with this formulation of Newton's method is that (15.4) is simpler to solve than (15.2). In fact, various quadratic programming techniques can be used to solve (15.4): active-set strategies, interior-point methods, dual approaches, etc. We also see that the combinatorial aspect of the original problem, which lies in the determination of the active inequality constraints, is transferred to the QP (15.4), where it is simpler to deal with than in the original nonlinear problem. However, the SQP algorithm has its own cost, which should not be overlooked. Indeed, all constraints must be linearized, including the inactive inequalities, which should play no role when the iterates are close to a solution. If these are many, the algorithm may loose some efficiency. Careful implementations use techniques to deal more efficiently with this situation (see for example [324, 301]).

## Discarding Parasitic Displacements

The implementation of the SQP algorithm and the analysis of its local convergence are more complex than when only equality constraints are present. In fact, the quadratic problem (15.4) may be infeasible (its feasible set may be empty) or unbounded (the optimal value is $-\infty$ ), or it may have multiple local solutions (a nonconvexity effect), even in the neighborhood of a solution $\left(x_{*}, \lambda_{*}\right)$ to $\left(P_{E I}\right)$. This may happen even when $\left(x_{*}, \lambda_{*}\right)$ enjoys nice properties such as the second-order sufficient conditions of optimality, strict complementarity, and constraint qualification. Here is an example.

Example 15.1. We want to minimize the logarithm of $(1+x)$ for $x$ restricted to the interval $[0,3]$. In canonical form, the problem is

$$
\left\{\begin{array}{l}
\min _{x} \log (1+x) \\
-x \leq 0 \\
x-3 \leq 0
\end{array}\right.
$$

The logarithm has been used to introduce nonconvexity in the problem, since by the monotonicity of the logarithmic function, it is equivalent to minimize $(1+x)$ or $\log (1+x)$. It is easily checked that this problem has a unique primaldual solution $\left(x_{*}, \lambda_{*}\right)=(0,(1,0))$, which satisfies the second-order sufficient conditions of optimality, strict complementarity, and the constraint qualification (LI-CQ). It is therefore a "good" solution. However, the osculating QP (15.4) at this solution can be written

$$
\left\{\begin{array}{l}
\min _{d} d-\frac{1}{2} d^{2} \\
-d \leq 0 \\
-3+d \leq 0
\end{array}\right.
$$

This problem has three primal-dual stationary points $(d, \lambda)$ : a local minimum $(0,(1,0))$, a maximum $(1,(0,0))$ and a global minimum $(3,(0,2))$. It would be unbounded without the constraint $x \leq 3$ in the original problem, which
is inactive at the solution. Among these stationary points, only the first one is suitable: it gives a zero displacement (which is to be expected from an algorithm started at a solution!), and optimal multipliers. The other two stationary points are parasitic.

The situation of this example can only occur if $L_{k}$ is not positive definite. Otherwise, problem (15.4) is strictly convex and therefore has a unique solution as soon as the feasible set is nonempty. The convergence results given in $\S 15.2$ assume that the parasitic solutions to the QP, like those revealed in the example, are discarded. Specifically, this is done by assuming that $d_{k}$ is the minimum norm solution to the QP.

### 15.2 Primal-Dual Quadratic Convergence

We first analyze the well-posedness of the SQP algorithm and the convergence of the generated primal-dual sequences, when the first iterate is chosen in some neighborhood of a "regular" stationary point (a notion that is made precise in the statement of theorem 15.2 below) that satisfies strict complementarity. At such a stationary point, (LI-CQ) holds.

Theorem 15.2 highlights an interesting property of the SQP algorithm: in some neighborhood of a stationary point satisfying the assumptions above, the active constraints of the osculating quadratic problem (15.4) are the same as those of $\left(P_{E I}\right)$. We have said that the identification of the active constraints is a major difficulty when solving inequality constrained problems and that, in the SQP algorithm, this difficulty is transferred to the osculating quadratic problem (QP), where it is easier to deal with. The result below tells us more: the active constraints of an osculating QP at one iteration are likely to be the same at the next iteration, at least close to a regular stationary point. Numerically, this means that, at least asymptotically, it is advantageous to solve the osculating QP's by algorithms that can take advantage of a good guess of the active constraints. Then, the combinatorial problem of determining which are the active constraints at the solution no longer occurs during the last iterations of the SQP algorithm.

Observe that, as this was already the case for equality constrained problems, the SQP algorithm may well generate a sequence that converges to a stationary point of $\left(P_{E I}\right)$ that is not a minimum point of the problem. Observe indeed that, at any stationary point $\left(x_{*}, \lambda_{*}\right)$ of $\left(P_{E I}\right),\left(0, \lambda_{*}\right)$ is a primal-dual solution to the quadratic problem, so that the SQP algorithm suggests not leaving $x_{*}$. This is due to the fact that SQP has been designed by linearizing the optimality conditions and therefore the algorithm makes no distinction between minima, maxima, or other stationary points.

Theorem 15.2 (primal-dual quadratic convergence of the SQP algorithm). Suppose that $f$ and $c$ are of class $C^{2}$ in a neighborhood of a
stationary point $x_{*}$ of $\left(P_{E I}\right)$, with associated multiplier $\lambda_{*}$. Suppose also that strict complementarity holds and that $\left(x_{*},\left(\lambda_{*}\right)_{E \cup I_{*}^{0}}\right)$ is a regular stationary point of the equality constrained problem

$$
\left\{\begin{array}{l}
\min _{x} f(x)  \tag{15.5}\\
c_{i}(x)=0, \quad \text { for } i \in E \cup I_{*}^{0},
\end{array}\right.
$$

in the sense of definition 14.2. Consider the $S Q P$ algorithm, in which $d_{k}$ is a minimum norm stationary point of the osculating quadratic problem (15.4). Then there is a neighborhood $V$ of $\left(x_{*}, \lambda_{*}\right)$ such that, if the first iterate $\left(x_{1}, \lambda_{1}\right) \in V:$
(i) the SQP algorithm is well defined and generates a sequence $\left\{\left(x_{k}, \lambda_{k}\right)\right\}$ that converges superlinearly to $\left(x_{*}, \lambda_{*}\right)$;
(ii) the active constraints of the osculating quadratic problem (15.4) are those of problem $\left(P_{E I}\right)$;
(iii) if, in addition, $f$ and c are of class $C^{2,1}$ in a neighborhood of $x_{*}$, the convergence of $\left\{\left(x_{k}, \lambda_{k}\right)\right\}$ is quadratic.

Proof. The idea of the proof is to show that, close to $\left(x_{*}, \lambda_{*}\right)$, the selected minimum norm stationary point of the osculating quadratic problem (15.4) and the primal-dual Newton step for (15.5) are identical. The result then follows from theorem 14.4.

Suppose that $(x, \lambda)$ is close to $\left(x_{*}, \lambda_{*}\right)$. Since $\left(x_{*},\left(\lambda_{*}\right)_{E \cup I_{*}^{0}}\right)$ is a regular stationary point of (15.5), $c_{E \cup I_{*}^{0}}^{\prime}\left(x_{*}\right)$ is surjective and the quadratic program in $\tilde{d}$

$$
\left\{\begin{array}{l}
\min _{\tilde{d}} \nabla f(x)^{\top} \tilde{d}+\frac{1}{2} \tilde{d}^{\top} L(x, \lambda) \tilde{d}  \tag{15.6}\\
c_{i}(x)+c_{i}^{\prime}(x) \cdot \tilde{d}=0, \quad \text { for } i \in E \cup I_{*}^{0}
\end{array}\right.
$$

has a unique primal-dual stationary point. We denoted it by $\left(\tilde{d}, \tilde{\lambda}_{E \cup I_{*}^{0}}\right)$ and form with $\tilde{\lambda}_{E \cup I_{*}^{0}}$ a vector $\tilde{\lambda} \in \mathbb{R}^{m}$, by setting $\tilde{\lambda}_{i}=0$ for $i \in I \backslash I_{*}^{0}$.

Let us show that $(\tilde{d}, \tilde{\lambda})$ is a stationary point of the osculating quadratic problem (15.4), if $(x, \lambda):=\left(x_{k}, \lambda_{k}\right)$ is in some neighborhood of $\left(x_{*}, \lambda_{*}\right)$. We only need to show that $c_{i}(x)+c_{i}^{\prime}(x) \cdot \tilde{d} \leq 0$ for $i \in I \backslash I_{*}^{0}$ and $\lambda_{i} \geq 0$ for $i \in I_{*}^{0}$. From theorem 14.4, $(x+\tilde{d}, \tilde{\lambda})$ is close to $\left(x_{*}, \lambda_{*}\right)$, when $(x, \lambda)$ is close to $\left(x_{*}, \lambda_{*}\right)$. Therefore, for $i \in I_{*}^{0}, \tilde{\lambda}_{i} \geq 0$, since $\left(\lambda_{*}\right)_{i}>0$ by strict complementarity. On the other hand, $\tilde{d}$ is small, so that $c_{i}(x)+c_{i}^{\prime}(x) \cdot \tilde{d} \leq 0$ for $i \in I \backslash I_{*}^{0}$. Hence $(\tilde{d}, \tilde{\lambda})$ is a stationary point of (15.4). We deduce from this that, for $(x, \lambda)$ close to $\left(x_{*}, \lambda_{*}\right)$, the SQP algorithm is well defined and $d$ is small (it is a minimum norm stationary point and $\tilde{d}$ is small by theorem 14.4).

Let us now show that the pair $\left(d, \lambda^{\mathrm{QP}}\right):=\left(d_{k}, \lambda_{k}^{\mathrm{QP}}\right)$ formed of the minimum norm solution to the QP and its associated multiplier is in fact $(\tilde{d}, \tilde{\lambda})$, if $(x, \lambda)$ is in some neighborhood of $\left(x_{*}, \lambda_{*}\right)$. From theorem 14.4, this will conclude the proof. For $(x, \lambda)$ close to $\left(x_{*}, \lambda_{*}\right)$ and $i \in I \backslash I_{*}^{0}, c_{i}(x)+c_{i}^{\prime}(x) \cdot d<0$, so that $\lambda_{i}^{\text {QP }}=0=\tilde{\lambda}_{i}$. Because of the uniqueness of the stationary point of (15.6), it remains to show that $c_{i}(x)+c_{i}^{\prime}(x) \cdot d=0$ for all $i \in I_{*}^{0}$ and $(x, \lambda)$
close to $\left(x_{*}, \lambda_{*}\right)$. If this is not the case, there would exist an index $j \in I_{*}^{0}$ and a sequence $(x, \lambda) \rightarrow\left(x_{*}, \lambda_{*}\right)$, such that $c_{j}(x)+c_{j}^{\prime}(x) \cdot d<0$. Then $\lambda_{j}^{\mathrm{QP}}=0$ and

$$
\nabla f(x)+L(x, \lambda) d+\sum_{i \in\left(E \cup I_{*}^{0}\right) \backslash\{j\}} \lambda_{i}^{\mathrm{QP}} \nabla c_{i}(x)=0
$$

Since $\nabla f(x)+L(x, \lambda) d \rightarrow \nabla f\left(x_{*}\right)(d$ is smaller than $\tilde{d}$, which converges to 0$)$ and $c_{E \cup I_{*}^{0}}^{\prime}\left(x_{*}\right)$ is surjective, $\lambda_{i}^{\mathrm{QP}}$ for $i \in\left(E \cup I_{*}^{0}\right) \backslash\{j\}$ would converge to some limit, $\bar{\lambda}_{i}$ say. Taking the limit in the equation above would give

$$
\nabla f\left(x_{*}\right)+\sum_{i \in\left(E \cup I_{*}^{0}\right) \backslash\{j\}} \bar{\lambda}_{i} \nabla c_{i}\left(x_{*}\right)=0 .
$$

Therefore, we would have found two different multipliers: $\bar{\lambda}$ (we set $\bar{\lambda}_{i}=0$ for $\left.i \notin\left(E \cup I_{*}^{0}\right) \backslash\{j\}\right)$ and $\lambda_{*}\left(\bar{\lambda} \neq \lambda_{*}\right.$ since $\bar{\lambda}_{j}=0$ and $\left(\lambda_{*}\right)_{j}>0$ by strict complementarity). This would be in contradiction with the uniqueness of the multiplier, which follows from the surjectivity of $c_{E \cup I_{*}^{0}}^{\prime}\left(x_{*}\right)$.

It is clear from the proof of theorem 15.2 that it is not really necessary to take for $d_{k}$, a minimum norm stationary point of the osculating quadratic problem (15.4), some $d_{k}^{\min }$ say. The result is still true if the SQP algorithm ensures that $d_{k} \rightarrow 0$ when $d_{k}^{\min } \rightarrow 0$. For example, it would suffice to compute a stationary point $d_{k}$ satisfying an estimate of the form $\left\|d_{k}\right\| \leq C\left\|d_{k}^{\text {min }}\right\|$, for some positive constant $C$.

Theorem 15.4 below considers the case when strict complementarity does not hold, but assumes that $\left(x_{*}, \lambda_{*}\right)$ satisfies the second order sufficient conditions of optimality and linear independence of the active constraint gradients (LI-CQ). The result is also local, in the sense that the first iterate $\left(x_{1}, \lambda_{1}\right)$ is supposed to be close enough to $\left(x_{*}, \lambda_{*}\right)$. The proof of this result is more difficult. This is because one can no longer use theorem 14.4 as in the preceding proof: the SQP step may be different from the Newton step on (15.5), however close to $\left(x_{*}, \lambda_{*}\right)$ the current iterate $(x, \lambda)$ can be. In other words, the property of local identification of the active constraints by the osculating quadratic problem no longer holds when complementarity is not strict. Here is an example.

Example 15.3. Consider the problem in $x \in \mathbb{R}$ :

$$
\left\{\begin{array}{l}
\min _{x} x^{2}+x^{4} \\
x \leq 0
\end{array}\right.
$$

The solution is $x_{*}=0$ and $\lambda_{*}=0$, so that strict complementarity does not hold. On the other hand, the constraint is qualified at $x_{*}$ in the sense of (LI-CQ) and the second order sufficient conditions of optimality hold. The osculating quadratic problem at $x$ (it does not depend on $\lambda$ since the constraint is linear) is the problem in $d \in \mathbb{R}$ :

$$
\left\{\begin{array}{l}
\min _{d}\left(2 x+4 x^{3}\right) d+\left(1+6 x^{2}\right) d^{2} \\
x+d \leq 0
\end{array}\right.
$$

If $x>0, x+d=0$ and the solution is obtained in one step. But if $x<0$, $\left.x+d=4 x^{3} /\left(1+6 x^{2}\right) \in\right] 2 x / 3,0[$, so that the linearized constraint is inactive and the SQP step is different from the Newton step on (15.5). In this case, however, the convergence is cubic in $x$ (also in $(x, \lambda)$ ): $|x+d| /|x|^{3} \leq 4$.

The preceding example suggests that fast convergence can still be obtained even without strict complementarity. This is confirmed by the following theorem.

Theorem 15.4 (primal-dual quadratic convergence of the SQP algorithm). Suppose that $f$ and $c$ are of class $C^{2,1}$ in a neighborhood of a local solution $x_{*}$ to $\left(P_{E I}\right)$. Suppose also that the constraint qualification (LI$C Q)$ is satisfied at $x_{*}$ and denote by $\lambda_{*}$ the associated multiplier. Finally, suppose that the second-order sufficient condition of optimality (13.8) is satisfied. Consider the SQP algorithm, in which $d_{k}$ is a minimum norm stationary point of the osculating quadratic problem (15.4). Then there exists a neighborhood $V$ of $\left(x_{*}, \lambda_{*}\right)$ such that, if the first iterate $\left(x_{1}, \lambda_{1}\right) \in V$, the SQP algorithm is well defined and the sequence $\left\{\left(x_{k}, \lambda_{k}\right)\right\}$ converges quadratically to $\left(x_{*}, \lambda_{*}\right)$.

Proof. The following lemma is assumed (see [308]).
Lemma 15.5. Under the conditions of theorem 15.4, there exists a neighborhood of $\left(x_{*}, \lambda_{*}\right)$ such that (15.3) has a local solution and the local solution $\left(d_{k}, \lambda_{k}^{\mathrm{QP}}\right)$ with $d_{k}$ of minimum norm satisfies:

$$
\left\|d_{k}\right\|+\left\|\lambda_{k}^{\mathrm{QP}}-\lambda_{*}\right\| \leq C\left(\left\|x_{k}-x_{*}\right\|+\left\|\lambda_{k}-\lambda_{*}\right\|\right)
$$

From this lemma, the algorithm is well defined if ( $x_{k}, \lambda_{k}$ ) remains close to $\left(x_{*}, \lambda_{*}\right)$. This will result from the estimates obtained below.

Let us set

$$
\delta_{k}=\left\|x_{k}-x_{*}\right\|+\left\|\lambda_{k}-\lambda_{*}\right\| .
$$

From lemma 15.5, we have

$$
\begin{equation*}
d_{k}=O\left(\delta_{k}\right) \quad \text { and } \quad \lambda_{k+1}-\lambda_{*}=O\left(\delta_{k}\right) \tag{15.7}
\end{equation*}
$$

where $d_{k}$ is a minimum-norm solution to (15.4) and $\lambda_{k+1}=\lambda_{k}^{\mathrm{QP}}$ is the associated multiplier. We deduce that, for $i \in I \backslash I_{*}^{0}$ and $\delta_{k}$ small enough, we have

$$
c_{i}\left(x_{k}\right)+c_{i}^{\prime}\left(x_{k}\right) \cdot d_{k}<0
$$

Hence $\left(\lambda_{k+1}\right)_{i}=0$, and with the set of indices

$$
J=E \cup I_{*}^{0}
$$

the optimality of $d_{k}$ is expressed by

$$
L_{k} d_{k}+A_{J}\left(x_{k}\right)^{\top}\left(\lambda_{k+1}\right)_{J}+\nabla f_{k}=0 .
$$

A Taylor expansion of the left-hand side, using $\nabla_{x} \ell\left(x_{*}, \lambda_{*}\right)=0, x_{k+1}=$ $x_{k}+d_{k}$ and (15.7), leads to

$$
\begin{align*}
0 & =\nabla_{x} \ell\left(x_{k}, \lambda_{*}\right)+L\left(x_{k}, \lambda_{k}\right) d_{k}+A_{J}\left(x_{k}\right)^{\top}\left(\lambda_{k+1}-\lambda_{*}\right)_{J} \\
& =L_{*}\left(x_{k+1}-x_{*}\right)+A_{J}\left(x_{*}\right)^{\top}\left(\lambda_{k+1}-\lambda_{*}\right)_{J}+O\left(\delta_{k}^{2}\right) \tag{15.8}
\end{align*}
$$

Expand likewise the constraints of the osculating quadratic problem: we have for $i \in J$

$$
\begin{equation*}
c_{i}\left(x_{k}\right)+c_{i}^{\prime}\left(x_{k}\right) \cdot d_{k}=c_{i}^{\prime}\left(x_{*}\right) \cdot\left(x_{k+1}-x_{*}\right)+\left(\gamma_{k}\right)_{i}, \tag{15.9}
\end{equation*}
$$

where $\left(\gamma_{k}\right)_{i}=O\left(\delta_{k}^{2}\right)$.
From the assumption, $A_{J}\left(x_{*}\right)$ is surjective, so we can find a vector $v_{k} \in \mathbb{R}^{m}$ such that

$$
A_{J}\left(x_{*}\right) v_{k}=\left(\gamma_{k}\right)_{J} \quad \text { and } \quad v_{k}=O\left(\delta_{k}^{2}\right) .
$$

The last estimate can be obtained by taking a minimum-norm $v_{k}$ satisfying the first equation. With the notation

$$
w_{k}=x_{k+1}-x_{*}+v_{k},
$$

(15.9) becomes for $i \in J$ :

$$
\begin{equation*}
c_{i}\left(x_{k}\right)+c_{i}^{\prime}\left(x_{k}\right) \cdot d_{k}=c_{i}^{\prime}\left(x_{*}\right) \cdot w_{k} . \tag{15.10}
\end{equation*}
$$

The complementarity conditions of the osculating quadratic problem can be written

$$
\begin{equation*}
\left(\lambda_{k+1}\right)_{i}\left(c_{i}\left(x_{k}\right)+c_{i}^{\prime}\left(x_{k}\right) \cdot d_{k}\right)=0, \quad \text { for all } i \in I . \tag{15.11}
\end{equation*}
$$

Hence, if $\left(\lambda_{*}\right)_{i}>0$ and $\delta_{k}$ small enough, we have $c_{i}\left(x_{k}\right)+c_{i}^{\prime}\left(x_{k}\right) \cdot d_{k}=0$. Then we obtain from (15.10)

$$
\left\{\begin{array}{l}
c_{i}^{\prime}\left(x_{*}\right) \cdot w_{k}=0 \text { if } i \in E \cup I_{*}^{0+}  \tag{15.12}\\
c_{i}^{\prime}\left(x_{*}\right) \cdot w_{k} \leq 0 \text { if } i \in I_{*}^{00} .
\end{array}\right.
$$

This shows that $w_{k}$ lies in the critical cone $C_{*}$, defined by (13.6). From the second-order sufficiency condition, we then have for a constant $C_{1}>0$ :

$$
\begin{equation*}
C_{1}\left\|w_{k}\right\|^{2} \leq w_{k}^{\top} L_{*} w_{k} . \tag{15.13}
\end{equation*}
$$

Now compute $w_{k}^{\top} L_{*} w_{k}$. From (15.8) and $v_{k}=O\left(\delta_{k}^{2}\right)$,

$$
w_{k}^{\top} L_{*} w_{k}=-\left(\lambda_{k+1}-\lambda_{*}\right)_{J}^{\top} A_{J}\left(x_{*}\right) w_{k}+O\left(\left\|w_{k}\right\| \delta_{k}^{2}\right) \leq C_{2}\left\|w_{k}\right\| \delta_{k}^{2}
$$

since $\left(\lambda_{k+1}-\lambda_{*}\right)_{J}^{\top} A_{J}\left(x_{*}\right) w_{k}=0$ thanks to (15.11) and (15.12). With (15.13), we then obtain

$$
C_{1}\left\|w_{k}\right\| \leq C_{2} \delta_{k}^{2}
$$

Since $v_{k}=O\left(\delta_{k}^{2}\right)$, we deduce

$$
x_{k+1}-x_{*}=O\left(\delta_{k}^{2}\right)
$$

On the other hand, this estimate, (15.8) and the injectivity of $A_{J}\left(x_{*}\right)^{\top}$ show that

$$
\left(\lambda_{k+1}-\lambda_{*}\right)_{J}=O\left(\delta_{k}^{2}\right)
$$

Since $\left(\lambda_{k+1}\right)_{i}=\left(\lambda_{*}\right)_{i}=0$ for $i \in I \backslash I_{*}^{0}$, these last two estimates show the quadratic convergence of the sequence $\left\{\left(x_{k}, \lambda_{k}\right)\right\}$.

### 15.3 Primal Superlinear Convergence

Theorem 15.4 gives conditions for the quadratic convergence of $\left\{\left(x_{k}, \lambda_{k}\right)\right\}$. Actually, this implies neither quadratic nor superlinear convergence for $\left\{x_{k}\right\}$ (see exercise 14.8). Nevertheless, the following result (theorem 15.7) shows that, for the SQP algorithm using the Hessian of the Lagrangian in the quadratic programs (15.4), the sequence $\left\{x_{k}\right\}$ converges superlinearly. This result is interesting because it is often desirable to have fast convergence of this sequence.

We consider for this an algorithm slightly more general than the one described in $\S 15.1$, which encompasses the quasi-Newton versions of the method. We suppose that $\left\{x_{k}\right\}$ is generated by

$$
x_{k+1}=x_{k}+d_{k},
$$

where $d_{k}$ is a stationary point of the quadratic problem

$$
\left\{\begin{array}{l}
\min _{d} \nabla f\left(x_{k}\right)^{\top} d+\frac{1}{2} d^{\top} M_{k} d  \tag{15.14}\\
\left(c\left(x_{k}\right)+A\left(x_{k}\right) d\right)^{\#}=0
\end{array}\right.
$$

This is the same problem as (15.4), but the Hessian of the Lagrangian $L_{k}$ is replaced by a symmetric matrix $M_{k}$. Incidentally, note that the multiplier $\lambda_{k}$ is no longer explicitly used in the algorithm. Theorem 15.7 gives a necessary and sufficient condition on $M_{k}$ to guarantee superlinear convergence of $\left\{x_{k}\right\}$.

The optimality conditions of (15.14) are $\left(\lambda_{k}^{\mathrm{QP}}\right.$ is the multiplier associated with the constraints):

$$
\left\{\begin{array}{l}
(a) \nabla f_{k}+M_{k} d_{k}+A_{k}^{\top} \lambda_{k}^{Q P}=0 \\
(b)\left(c_{k}+A_{k} d_{k}\right)^{\#}=0  \tag{15.15}\\
(c)\left(\lambda_{k}^{\mathrm{QP}}\right)_{I} \geq 0 \\
(d)\left(\lambda_{k}^{\mathrm{QP}}\right)_{I}\left(c_{k}+A_{k} d_{k}\right)_{I}=0
\end{array}\right.
$$

We shall need the orthogonal projector onto the critical cone $C_{*}$ at a solution $x_{*}$ to $\left(P_{E I}\right)$ (see (13.6)). We denote this (nonlinear) projector by $P_{*}$. It is well defined since $C_{*}$ is a nonempty closed convex set.

Lemma 15.6. If $\lambda \in \mathbb{R}^{m}$ is such that $\lambda_{I_{*}^{00}} \geq 0$ and $\lambda_{I \backslash I_{*}^{0}}=0$, then $P_{*} A_{*}^{\top} \lambda=0$.

Proof. Take $\lambda \in \mathbb{R}^{m}$ as in the terms of the lemma and $h \in C_{*}$. Then $\left(A_{*} h\right)_{E \cup I_{*}^{0+}}=0,\left(A_{*} h\right)_{I_{*}^{00}} \leq 0$, and we have

$$
\left(0-A_{*}^{\top} \lambda\right)^{\top}(h-0)=-\lambda^{\top} A_{*} h=-\lambda_{I_{*}^{00}}^{\top}\left(A_{*} h\right)_{I_{*}^{00}} \geq 0
$$

The characterization (13.12) of the projection yields the result.
Theorem 15.7 (primal superlinear convergence of the SQP algorithm). Suppose that $f$ and $c$ are twice differentiable at $x_{*} \in \Omega$. Suppose also that $\left(x_{*}, \lambda_{*}\right)$ is a primal-dual solution to $\left(P_{E I}\right)$ satisfying (LI-CQ) and the second-order sufficient condition of optimality (13.8). Consider the sequence $\left\{\left(x_{*}, \lambda_{*}\right)\right\}$ generated by the recurrence $x_{k+1}=x_{k}+d_{k}$ and $\lambda_{k+1}=\lambda_{k}^{\mathrm{QP}}$, where $\left(d_{k}, \lambda_{k}^{\mathrm{QP}}\right)$ is a primal-dual solution to (15.14). Suppose that $\left\{\left(x_{k}, \lambda_{k}\right)\right\}$ converges to $\left(x_{*}, \lambda_{*}\right)$. Then $\left\{x_{k}\right\}$ converges superlinearly if and only if

$$
\begin{equation*}
P_{*}\left(L_{*}-M_{k}\right) d_{k}=o\left(\left\|d_{k}\right\|\right) \tag{15.16}
\end{equation*}
$$

where $P_{*}$ is the orthogonal projector onto the critical cone $C_{*}$.
Proof. Using (15.15) $)_{a}, \nabla_{x} \ell\left(x_{*}, \lambda_{*}\right)=0$ and $\lambda_{k+1} \rightarrow \lambda_{*}$, we have

$$
\begin{aligned}
-M_{k} d_{k} & =\nabla_{x} \ell\left(x_{k}, \lambda_{k+1}\right) \\
& =\nabla_{x} \ell\left(x_{*}, \lambda_{k+1}\right)+L\left(x_{*}, \lambda_{k+1}\right)\left(x_{k}-x_{*}\right)+o\left(\left\|x_{k}-x_{*}\right\|\right) \\
& =A_{*}^{\top}\left(\lambda_{k+1}-\lambda_{*}\right)+L_{*}\left(x_{k}-x_{*}\right)+o\left(\left\|x_{k}-x_{*}\right\|\right)
\end{aligned}
$$

Hence

$$
\begin{equation*}
\left(L_{*}-M_{k}\right) d_{k}=A_{*}^{\top}\left(\lambda_{k+1}-\lambda_{*}\right)+L_{*}\left(x_{k+1}-x_{*}\right)+o\left(\left\|x_{k}-x_{*}\right\|\right) \tag{15.17}
\end{equation*}
$$

To show that condition (15.16) is necessary, assume that $x_{k+1}-x_{*}=$ $o\left(\left\|x_{k}-x_{*}\right\|\right)$. Then (15.17) gives

$$
\left(L_{*}-M_{k}\right) d_{k}=A_{k}^{\top}\left(\lambda_{k+1}-\lambda_{*}\right)+o\left(\left\|x_{k}-x_{*}\right\|\right)
$$

Project with $P_{*}$, which is Lipschitzian (see (13.15)), and observe that, from $(15.15)_{c}$ and $(15.15)_{d},\left(\lambda_{k+1}-\lambda_{*}\right)$ satisfies for large $k$ the conditions on $\lambda$ of lemma 15.6:

$$
P_{*}\left(L_{*}-M_{k}\right) d_{k}=P_{*} A_{*}^{\top}\left(\lambda_{k+1}-\lambda_{*}\right)+o\left(\left\|x_{k}-x_{*}\right\|\right)=o\left(\left\|x_{k}-x_{*}\right\|\right)
$$

Condition (15.16) follows, because $\left(x_{k}-x_{*}\right) \sim d_{k}$ by lemma 13.5.
Conversely, let us show that condition (15.16) is sufficient. For $i \in J:=$ $E \cup I_{*}^{0}$, we have

$$
c_{i}\left(x_{k}\right)+c_{i}^{\prime}\left(x_{k}\right) \cdot d_{k}=c_{i}^{\prime}\left(x_{*}\right) \cdot\left(x_{k+1}-x_{*}\right)+\left(\gamma_{k}\right)_{i}
$$

where $\left(\gamma_{k}\right)_{i}=o\left(\left\|x_{k}-x_{*}\right\|\right)+o\left(\left\|d_{k}\right\|\right)$. Since $A_{J}\left(x_{*}\right)$ is surjective, $\left(\gamma_{k}\right)_{J}=$ $A_{J}\left(x_{*}\right) v_{k}$, for some $v_{k}=o\left(\left\|x_{k}-x_{*}\right\|\right)+o\left(\left\|d_{k}\right\|\right)$. With the notation

$$
w_{k}:=x_{k+1}-x_{*}+v_{k},
$$

there holds

$$
c_{i}\left(x_{k}\right)+c_{i}^{\prime}\left(x_{k}\right) \cdot d_{k}=c_{i}^{\prime}\left(x_{*}\right) \cdot w_{k}, \quad \text { for } i \in J
$$

Now $c_{i}\left(x_{k}\right)+c_{i}^{\prime}\left(x_{k}\right) \cdot d_{k}=0$ for $i \in E \cup I_{*}^{0+}$ and $k$ large enough, so that

$$
\left\{\begin{array}{l}
c_{i}^{\prime}\left(x_{*}\right) \cdot w_{k}=0 \text { if } i \in E \cup I_{*}^{0+} \\
c_{i}^{\prime}\left(x_{*}\right) \cdot w_{k} \leq 0 \text { if } i \in I_{*}^{00} .
\end{array}\right.
$$

This implies that $w_{k} \in C_{*}$ for large $k$ (see (13.6)) and that, for some constant $C_{1}>0$,

$$
\begin{equation*}
C_{1}\left\|w_{k}\right\|^{2} \leq w_{k}^{\top} L_{*} w_{k}, \quad \text { for large } k \tag{15.18}
\end{equation*}
$$

On the other hand, for $i \in I_{*}^{00}$, from (15.15) , we have $0=\left(\lambda_{k+1}\right)_{i}\left(c_{i}\left(x_{k}\right)+\right.$ $\left.c_{i}^{\prime}\left(x_{k}\right) \cdot d_{k}\right)=\left(\lambda_{k+1}\right)_{i}\left(c_{i}^{\prime}\left(x_{*}\right) \cdot w_{k}\right)$ and $\left(\lambda_{*}\right)_{i}=0$. While for $i \in I \backslash I_{*}^{0}$, $\left(\lambda_{k+1}-\lambda_{*}\right)_{i}=0$. Therefore

$$
\left(\lambda_{k+1}-\lambda_{*}\right)^{\top} A_{*} w_{k}=0, \quad \text { for large } k
$$

Now, with this equation, (15.17), $v_{k}=o\left(\left\|x_{k}-x_{*}\right\|\right)+o\left(\left\|d_{k}\right\|\right)$, the fact that $u^{\top} v \leq u^{\top} P_{*} v$, for all $v \in \mathbb{R}^{n}$ and all $u \in C_{*}$ (see (13.14)), and (15.16), we find that

$$
\begin{aligned}
w_{k}^{\top} L_{*} w_{k} & =w_{k}^{\top} L_{*}\left(x_{k+1}-x_{*}\right)+O\left(\left\|w_{k}\right\|\left\|v_{k}\right\|\right) \\
& =w_{k}^{\top}\left(L_{*}-M_{k}\right) d_{k}+o\left(\left\|w_{k}\right\|\left\|x_{k}-x_{*}\right\|\right)+o\left(\left\|w_{k}\right\|\left\|d_{k}\right\|\right) \\
& \leq w_{k}^{\top} P_{*}\left(L_{*}-M_{k}\right) d_{k}+o\left(\left\|w_{k}\right\|\left\|x_{k}-x_{*}\right\|\right)+o\left(\left\|w_{k}\right\|\left\|d_{k}\right\|\right) \\
& =o\left(\left\|w_{k}\right\|\left\|x_{k}-x_{*}\right\|\right)+o\left(\left\|w_{k}\right\|\left\|d_{k}\right\|\right) .
\end{aligned}
$$

With (15.18), $w_{k}=o\left(\left\|x_{k}-x_{*}\right\|\right)+o\left(\left\|d_{k}\right\|\right)$; hence

$$
x_{k+1}-x_{*}=o\left(\left\|x_{k}-x_{*}\right\|\right)+o\left(\left\|d_{k}\right\|\right) .
$$

The property $x_{k+1}-x_{*}=o\left(\left\|x_{k}-x_{*}\right\|\right)$ follows easily.
When there are no inequality constraints, $P_{*}$ is the orthogonal projector onto the null space $N\left(A_{*}\right)$. It is then linear. Given a basis $Z_{*}^{-}$of $N\left(A_{*}\right)$, it can be written

$$
P_{*}=Z_{*}^{-}\left(Z_{*}^{-\top} Z_{*}^{-}\right)^{-1} Z_{*}^{-\top} .
$$

Since $Z_{*}^{-}$is injective and $Z_{*}^{-\top} Z_{*}^{-}$is nonsingular, condition (15.16) can be written

$$
Z_{*}^{-\top}\left(L_{*}-M_{k}\right) d_{k}=o\left(\left\|d_{k}\right\|\right) \quad \text { or } \quad\left(Z_{*}^{-\top} L_{*}-Z_{k}^{-\top} M_{k}\right) d_{k}=o\left(\left\|d_{k}\right\|\right) .
$$

To write the last condition, we have supposed that $Z^{-}(\cdot)$ is continuous at $x_{*}$ and that $\left\{M_{k}\right\}$ is bounded. This shows that the important part of $M_{k}$ is $Z_{k}^{-\top} M_{k}$, which reminds us that only the part $Z_{k}^{-\top} L_{k}$ of $L_{k}$ plays a role in the definition of the Newton direction for equality constrained problems (see observation 1 on page 235).

### 15.4 The Hanging Chain Project III

In this third session, we resume the project on the determination of the static equilibrium position of a hanging chain, started in $\S 13.8$ and developed in $\S$ 14.7. Our present objective is to implement the local SQP algorithm, presented on page 257 , to be able to take into account the floor constraint. The algorithm is quite similar to the Newton method implemented in the second session. The main difference is that the solver of linear equations has to be replaced by a solver of quadratic optimization problems. This simple change will have several consequences that are discussed in this section.

It is a good idea to keep the work done in the second session and to use $\mathrm{mi}=m_{I}$ as a flag that makes the sqp function select the type of solver (linear or quadratic), depending on the presence of inequality constraints. Solving a linear system is indeed much simpler than solving a quadratic optimization problem, so that the sqp function must be allowed to take advantage of the absence of inequality constraints.

## Modifications to Bring to the sqp Function

Most of the work has been done in the previous session. There are only two modifications to bring to the function sqp.

The main change consists in substituting a quadratic optimization solver (to solve (15.4)) for the linear solver previously used in sqp (see chapter 14). Writing a solver of quadratic optimization problems is a difficult task. Fortunately, in our case, the Matlab solver quadprog can be used, so that we can concentrate on other aspects of the SQP algorithm. Quadprog first finds an initial feasible point by solving a linear optimization problem and then uses an active set method to find a solution to the quadratic problem. It can detect infeasibility and unboundedness.

A second change deals with the determination of the initial dual solution $\lambda=\left(\lambda_{E}, \lambda_{I}\right)$. Since it is known that $\lambda_{I}$ must be nonnegative, it is better now to determine $\lambda$ as a solution to the bound constrained least-squares problem

$$
\min _{\substack{\lambda=\left(\lambda_{E}, \lambda_{I}\right) \in \mathbb{R}^{m} \\ \lambda_{I} \geq 0}} \frac{1}{2}\left\|\nabla_{x} \ell(x, \lambda)\right\|_{2}^{2}
$$

instead of using (14.47). This convex quadratic optimization problem always has a solution (theorem 19.1). It can be solved by quadprog.

## Checking the Correctness of the SQP Solver

There is little change to make an error on the part of the simulator dealing with the inequality constraints, since these are very simple. Nevertheless, it is better to check it and to verify the implementation of the quadratic solver.

The same strategy as in the case with equality constrained problems can be followed: trying to solve more and more difficult problems and check the quadratic convergence of the generated sequence.

Let us check the quadratic convergence on the following variant of test case 1a, in which we add a floor constraint.

Test case 1e: same data as for the test case 1a (namely second hook at $(a, b)=(1,-0.3)$ and bars of lengths $L=(0.4,0.3,0.25,0.2,0.4))$ with an additional floor with parameters $\left(g_{0}, g_{1}\right)=(-0.35,-0.2)$ (see the definition of the floor in (13.25)). The initial positions of the chain joints are $(0.1,-0.3)$, $(0.4,-0.5),(0.6,-0.4)$, and $(0.7,-0.5)$.
The results obtained with test case 1 e are shown in figure 15.2. Convergence


Fig. 15.2. Test case 1 e
with options.tol $(1: 4)=10^{-10}$ is again obtained in 6 iterations. The picture on the left uses the same conventions as before: the thin solid bars represent the initial position of the chain, the dashed bars correspond to the 5 intermediate positions (hardly distinguishable), and the bold solid bars are those of the final optimal position. This one is a local minimum (the multipliers associated with the inequality constraints are positive and the critical cone is reduced to $\{0\})$. The picture on the right gives a plot of the ratios $\left\|s_{k+1}\right\|_{2} /\left\|s_{k}\right\|_{2}^{2}$, where $s_{k}=z_{k+1}-z_{k}$, for $k=1, \ldots, 5$. The boundedness of these ratios shows without any doubt that the sequence $\left\{z_{k}\right\}=\left\{\left(x_{k}, \lambda_{k}\right)\right\}$ converges quadratically to its limit, as predicted by the theory (theorems 15.2 and 15.4).

## Experimenting with the SQP Algorithm

A first observation, with unpleasant consequences, is that quadprog is aimed at computing a local minimum of a quadratic problem, not an arbitrary stationary point (note that finding a solution or a stationary point of a nonconvex quadratic problem is NP-hard, see [354] for example). Therefore, it is quite frequent to find situations where quadprog fails to find a stationary
point, as required by the SQP algorithm. For example, with test case $1 f$ below, which is test case 1 e with the initial position of the chain given in test case 1a, the first osculating quadratic problem is unbounded.
Test case 1f: same data as for the test case 1e; but the initial positions of the chain joints are $(0.2,-0.5),(0.4,-0.6),(0.6,-0.8)$, and $(0.8,-0.6)$.
The unboundedness of an osculating quadratic problem can occur only when its feasible set is unbounded and the Hessian of the Lagrangian $L$ is not positive definite at the current iterate. Hence, taking a positive definite approximation of $L$ cures the difficulty. This can be obtained by using a quasi-Newton approximation of $L$; this technique is considered in chapter 18. Another possibility is add to $L$ a (not too large) positive diagonal matrix $E$, such that $L+E$ is positive definite (for example by using a modified Cholesky factorization of $L$ [154, 201]). Figure 15.3 shows the results obtained with this technique.


Fig. 15.3. Test case 1f

The optimal chain is actually a local minimum (the critical cone is reduced to $\{0\}$ and the energy is $e=-0.489$ ), different from the one obtained in figure 15.2 (in which $e=-0.518$ ). Observe that, although the initial position of the chain is not feasible for the floor constraint, the subsequent positions are all feasible. This is due to the affinity of the floor constraint (see (13.25) and exercise 15.1).

Another difficulty arises when the linearized constraints are incompatible, leading to an infeasible osculating quadratic problem. This difficulty is encountered at the first iteration with the initial chain given in test case 1 g below. Remedies for this kind of situations exist, see [351, 341, 156] and the references thereof.

Test case 1 g : same data as for the test case 1 e ; but the initial positions of the chain joints are $(0.1,-0.3),(0.3,-0.4),(0.6,-0.4)$, and $(0.7,-0.4)$.

## Notes

The SQP algorithm, in a form equivalent to the one introduced in $\S 15.1$ on page 257 , was first proposed by Wilson [359; 1963]. This author was mainly concerned with the extension of the simplex method, first to quadratic programming, and then to nonlinear convex optimization problems. The algorithm was obtained by searching for a saddle point of a quadratic approximation of the Lagrangian in the primal and dual variables. No convergence proof was given. See also the introduction of this part of the book, on page 191, for other references on the origin of the SQP algorithm.

The local quadratic convergence of theorem 15.2 is due to several authors; see for example [307], in which various classes of algorithms are considered. Theorem 15.4 is taken from [38]; further refinements can be found in [40].

The criterion (15.16) for superlinear convergence dates back to Dennis and Moré [104], who introduced a similar condition to characterize the superlinear convergence of sequences generated by quasi-Newton methods in unconstrained optimization (see theorem 4.11). It was extended to problems with equality constraints by Boggs, Tolle, and Wang [36], under a somewhat strong assumption (linear convergence of the sequence $\left\{x_{k}\right\}$ ). The possibility of getting rid of this assumption has been observed by many authors. The generalization to inequality constrained problems given in theorem 15.7 is due to Bonnans [40], who uses a projector varying along the iterations; in contrast, we use the projector $P_{*}$ onto the critical cone.

The local convergence of the SQP algorithm has been extended to different contexts, such as semi-infinite programming [180], infinite dimension programming [3, 4, 5, 6, 219]. When (MF-CQ) holds, but not (LI-CQ), the optimal multiplier may not be unique, so that the limit behavior of the multiplier sequence $\left\{\lambda_{k}^{\mathrm{QP}}\right\}$ is difficult to predict; this situation is analyzed in $[367,183,7,8]$.

## Exercise

15.1. Consider the SQP algorithm applied to problem $\left(P_{E I}\right)$ in which the $i$ th constraint, for some $i \in E \cup I$ (equality or inequality constraint), is affine (i.e., $c_{i}(x+d)=c_{i}(x)+c_{i}^{\prime}(x) \cdot d$ for all $x$ and $d \in \mathbb{R}^{n}$ ). Let $(x, \lambda)$ be the current iterate and define $x_{+}$by $x_{+}:=x+\alpha d$, where $d$ is a solution to the osculating quadratic problem (15.4) (we drop the index $k$ ) and $\alpha \in] 0,1]$. Show that $x_{+}$is feasible for the $i$ th constraint (i.e., $c_{i}\left(x_{+}\right)=0$ if $i \in E$, or $c_{i}\left(x_{+}\right) \leq 0$ if $\left.i \in I\right)$ if either $x$ is feasible for the $i$ th constraint or if $\alpha=1$.

## 16 Exact Penalization

### 16.1 Overview

The algorithms studied in chapters 14 and 15 generate converging sequences if the first iterate is close enough to a regular stationary point (see theorems $14.4,14.5,14.7,15.2$, and 15.4 ). Such an iterate is not necessarily at hand, so it is important to have techniques that allow the algorithms to force convergence, even when the starting point is far from a solution. This is known as the globalization of a local algorithm. The term is a little ambiguous, since it may suggest that it has a link with the search of global minimizers of $\left(P_{E I}\right)$. This is not at all the case (for an entry point on global optimization, see [200]).

There are (at least) two classes of techniques to globalize a local algorithm: line-search and trust-region; we shall only consider the line-search approach in this survey. Both techniques use the same idea: the progress made from one iterate $x_{k}$ to the next one $x_{k+1}$ towards the solution is measured by means of an auxiliary function, called the merit function (the novel notion of filter, not discussed in this part, looks like a promising alternative; see [130] for the original paper). In unconstrained optimization, "the" appropriate merit function is of course the objective $f$ itself. Here, the measure has to take into account the two, usually contradictory, goals in $\left(P_{E I}\right)$ : minimizing $f$ and satisfying the constraints. Accordingly, the merit function has often the following form

$$
f(x)+p(x)
$$

where $p$ is a function penalizing the constraint violation: $p$ is zero on the feasible set and positive outside. Instead of merit functions, one also speaks of penalty functions, although the latter term is usually employed when the penalty function is minimized by algorithms for unconstrained optimization. As we shall see, the approach presented here is more subtle: truly constrained optimization algorithms are used (like those in chapters 14 and 15); the merit function only intervenes as a tool for measuring the adequacy of the step computed by the local methods. It is not used for computing the direction itself. The main advantage is that the ill-conditioning encountered with penalty methods is avoided, and the fast speed of convergence of the local methods is ensured close to a solution.

As many merit functions exist, a selection must be made. We shall only study those that do not use the derivatives of $f$ and $c$. These are the most widely encountered in optimization codes and their numerical effectiveness has been demonstrated. To start with, let us examine some common examples of merit/penalty functions. We denote by $\|\cdot\|_{2}$ the $\ell_{2}$ norm and by $\|\cdot\|_{P}$ an arbitrary norm.
(a) Quadratic penalization:

$$
\begin{equation*}
f(x)+\frac{\sigma}{2}\left\|c(x)^{\#}\right\|_{2}^{2} \tag{16.1}
\end{equation*}
$$

(b) Lagrangian:

$$
f(x)+\mu^{\top} c(x)
$$

(c) Augmented Lagrangian (case $I=\emptyset$ ):

$$
\begin{equation*}
f(x)+\mu^{\top} c(x)+\frac{\sigma}{2}\|c(x)\|_{2}^{2} . \tag{16.2}
\end{equation*}
$$

Augmented Lagrangian (general case):

$$
\begin{align*}
f(x) & +\mu_{E}^{\top} c_{E}(x)+\frac{\sigma}{2}\left\|c_{E}(x)\right\|_{2}^{2} \\
& +\sum_{i \in I}\left(\mu_{i} \max \left(\frac{-\mu_{i}}{\sigma}, c_{i}(x)\right)+\frac{\sigma}{2}\left[\max \left(\frac{-\mu_{i}}{\sigma}, c_{i}(x)\right)\right]^{2}\right) \tag{16.3}
\end{align*}
$$

(d) Nondifferentiable augmented function:

$$
f(x)+\sigma\left\|c(x)^{\#}\right\|_{P}
$$

These functions have quite different features. One important property that distinguishes them is the exactness of the penalization, which is the subject of the present chapter. The concept of exact penalization is sometimes ambiguous - or at least varies from author to author. We adopt the following definition.

A function $\Theta: \Omega \rightarrow \mathbb{R}$ is called an exact penalty function at a local minimum $x_{*}$ of $\left(P_{E I}\right)$ if $x_{*}$ is a local minimum of $\Theta$. The converse implication $\left(x_{*}\right.$ is a local minimum of $\left(P_{E I}\right)$ whenever it minimizes $\Theta$ locally $)$ is not generally possible unless feasibility of $x_{*}$ is assumed. The example in figure 16.1 is an illustration: $x_{*}^{\prime}$ is a local minimum of the functions $(a)$ or $(d)$ with $\sigma \geq 0$ but, being infeasible, it is not a solution to the minimization problem. The reason why the concept of exactness is so important for globalizing the SQP algorithm will be discussed in chapter 17 .

Table 16.1 gives some properties of the merit functions $(a)-(d)$. This deserves some comments.

- As far as the differentiability of $\Theta_{\sigma}$ is concerned, we assume that $f$ and $c$ are of class $C^{\infty}$. We see that, in general, the presence of inequality constraints decreases the degree of differentiability of the merit functions. In this respect, the Lagrangian (b) is an exception.


Fig. 16.1. Exactness and feasibility

| Function | Differentiability Exactness | Conditions for <br> exactness | Threshold of $\sigma$ <br> depends on |  |
| :---: | :---: | :---: | :---: | :---: |
| $(a)$ | $C^{1}$ | no |  |  |
| $(b)$ | $C^{\infty}$ | yes | $\left(P_{E I}\right)$ convex <br> $\mu=\lambda_{*}$ |  |
| $(c)$ | $C^{1}$ | yes | $\mu=\lambda_{*}$ <br> $\sigma$ large | 2nd derivatives |
| $(d)$ | $C^{0}$ | yes | $\sigma$ large | 1st derivatives |

Table 16.1. Comparison of some merit functions

- We also see that only functions $(b)-(d)$ can be exact. The quadratic penalty function is hardly ever exact: if $I=\emptyset$, it is differentiable and its gradient at a solution $x_{*}$ is $\nabla f\left(x_{*}\right)$, which is usually nonzero. As we shall see in the following sections, the Lagrangian $(b)$ is exact for convex problems and the augmented Lagrangian (c) is exact for nonconvex problems provided the penalty parameter $\sigma$ is large enough.
- To be exact, both functions (b) and (c) need to have $\mu=\lambda_{*}$. From an algorithmic point of view, this means that the value of $\mu$ must be continually modified in order to approximate the unknown optimal multiplier $\lambda_{*}$. Algorithms using the Lagrangians do not minimize the same function at each iteration, which can raise convergence difficulties.
- Another shortcoming of $(c)$ is that the threshold of $\sigma$, beyond which the penalization becomes exact, involves the eigenvalues of the Hessian of the Lagrangian. It is therefore not easily accessible to computation, and certainly out of reach if the Hessian of the Lagrangian is not explicitly computed, as in the quasi-Newton versions of the algorithms. Nevertheless, many algorithms use this function (for example, those described in [85]).
- Finally, the conditions for the exactness of function $(d)$ are less restrictive and this is the main reason why this merit function is often used for globalizing the SQP algorithm, as in chapter 17 . We shall see in particular that the threshold of $\sigma$ can easily be estimated during the iterations, with the help of an estimate of the optimal multiplier. Function $(d)$ is nonsmooth, however.

In the rest of this chapter, we study some properties of the merit functions (b)-(d), focusing on their exactness.

### 16.2 The Lagrangian

In this section, problem $\left(P_{E I}\right)$ is assumed to be convex: $f$ and the $c_{i}$ 's, $i \in I$, are convex, and $c_{E}$ is affine. In this case, the Lagrangian of the problem is exact at a solution $x_{*}$, providing the multiplier is set to a dual solution $\lambda_{*}$. Actually, proposition 16.1 below shows a little more than that: $\ell$ has a saddlepoint at $\left(x_{*}, \lambda_{*}\right)$, a concept made precise in the next definition.

Let $X$ and $Y$ be two sets and let $\varphi: X \times Y \rightarrow \mathbb{R}$ be a function. We say that $\left(x_{*}, y_{*}\right) \in X \times Y$ is a saddle-point of $\varphi$ on $X \times Y$ when

$$
\varphi\left(x_{*}, y\right) \leq \varphi\left(x_{*}, y_{*}\right) \leq \varphi\left(x, y_{*}\right), \quad \text { for all } x \in X \text { and } y \in Y
$$

Thus, $x \mapsto \varphi\left(x, y_{*}\right)$ is minimal at $x_{*}$ and $y \mapsto \varphi\left(x_{*}, y\right)$ is maximal at $y_{*}$.
Recall that the Lagrangian of problem $\left(P_{E I}\right)$ is the function

$$
\begin{equation*}
(x, \mu) \in \Omega \times \mathbb{R}^{m} \mapsto \ell(x, \mu)=f(x)+\mu^{\top} c(x) \tag{16.4}
\end{equation*}
$$

If a feasible point $x_{*}$ minimizes $\ell(\cdot, \mu)$, then $0=\nabla_{x} \ell\left(x_{*}, \mu\right)$, which indicates that $x_{*}$ will be a solution to $\left(P_{E I}\right)$ provided $\mu$ is a dual solution. The following result shows that, for convex problems, the primal-dual solutions to $\left(P_{E I}\right)$ are saddle-points of $\ell$ on $\Omega \times\left\{\mu \in \mathbb{R}^{m}: \mu_{I} \geq 0\right\}$. The way is then open to computing primal-dual solutions to ( $P_{E I}$ ) with algorithms computing saddlepoints. We shall not proceed in that way but it is useful to bear this point of view in mind. In addition, this result shows that $x \mapsto \ell\left(x, \lambda_{*}\right)$ is an exact penalty function for convex problems.
Proposition 16.1 (saddle-point of the Lagrangian). Suppose that problem $\left(P_{E I}\right)$ is convex, that $x_{*}$ is a solution, and that $f$ and $c$ are differentiable at $x_{*}$. Suppose also that there exists a multiplier $\lambda_{*}$ such that the optimality conditions (KKT) are satisfied. Then $\left(x_{*}, \lambda_{*}\right)$ is a saddle-point of the Lagrangian defined in (16.4) on $\Omega \times\left\{\mu \in \mathbb{R}^{m}: \mu_{I} \geq 0\right\}$.
Proof. Take $\mu \in\left\{\mu \in \mathbb{R}^{m}: \mu_{I} \geq 0\right\}$. We have

$$
\begin{array}{rlr}
\ell\left(x_{*}, \mu\right) & =f\left(x_{*}\right)+\mu_{I}^{\top} c_{I}\left(x_{*}\right) \quad\left[\text { because } c_{E}\left(x_{*}\right)=0\right] \\
& \leq f\left(x_{*}\right) \quad\left[\text { because } \mu_{i} c_{i}\left(x_{*}\right) \leq 0 \text { for } i \in I\right] \\
& =f\left(x_{*}\right)+\lambda_{*}^{\top} c\left(x_{*}\right) \quad\left[\text { because } c_{E}\left(x_{*}\right)=0 \text { and }\left(\lambda_{*}\right)_{I}^{\top} c_{I}\left(x_{*}\right)=0\right] \\
& =\ell\left(x_{*}, \lambda_{*}\right) . &
\end{array}
$$

On the other hand, since $\left(\lambda_{*}\right)_{I} \geq 0$ and $\left(P_{E I}\right)$ is convex, the function $x \in \Omega \mapsto \ell\left(x, \lambda_{*}\right)$ is convex. According to the assumptions, this function is differentiable at $x_{*}$ and, in view of the optimality conditions (KKT), we have $\nabla_{x} \ell\left(x_{*}, \lambda_{*}\right)=0$. We deduce that this function is minimal at $x_{*}: \ell\left(x_{*}, \lambda_{*}\right) \leq$ $\ell\left(x, \lambda_{*}\right)$, for all $x \in \Omega$.

### 16.3 The Augmented Lagrangian

The Lagrangian (16.4) is not an exact penalty function if the problem is nonconvex. For example, the nonconvex problem

$$
\left\{\begin{array}{l}
\min _{x} \log (x) \\
x \geq 1
\end{array}\right.
$$

has the unique primal-dual solution $\left(x_{*}, \lambda_{*}\right)=(1,1)$ and its Lagrangian $\ell\left(x, \lambda_{*}\right)=\log (x)+1-x$ is concave with a maximum at $x=1$.

The augmented Lagrangian $\ell_{r}$ obviates this shortcoming. In fact we shall prove a local version of proposition 16.1: if $\mu=\lambda_{*}$ and $r$ is large enough, $\ell_{r}(\cdot, \mu)$ has a strict local minimum at a strong solution to the optimization problem $\left(P_{E I}\right)$.

The augmented Lagrangian is best introduced by using a perturbation technique as in duality theory, but this is beyond the scope of this review. Here we follow a more intuitive approach, starting with the case where only equality constraints are present. In this case, one takes

$$
\begin{equation*}
\ell_{r}(x, \mu)=f(x)+\mu_{E}^{\top} c_{E}(x)+\frac{r}{2}\left\|c_{E}(x)\right\|_{2}^{2} \tag{16.5}
\end{equation*}
$$

This is the standard Lagrangian $\ell$, augmented by the term $(r / 2)\left\|c_{E}(x)\right\|_{2}^{2}$. This term penalizes the constraint violation and makes $\ell_{r}(\cdot, \mu)$ convex around the point $x_{*}$ in a subspace complementary to the tangent space $N\left(A_{E}\left(x_{*}\right)\right)$. This creates a basin around a strong solution to $\left(P_{E}\right)$, making the penalization exact (this point of view is developed in exercise 16.2).

To deal with inequality constraints, we first transform $\left(P_{E I}\right)$ by introducing slack variables $s \in \mathbb{R}^{m_{I}}$ :

$$
\left\{\begin{array}{l}
\min _{(x, s)} f(x) \\
c_{E}(x)=0 \\
c_{I}(x)+s=0 \\
s \geq 0
\end{array}\right.
$$

Next, this problem is approached by using the augmented Lagrangian associated with its equality constraints:

$$
\min _{x} \min _{s \geq 0}\left(f(x)+\mu_{E}^{\top} c_{E}(x)+\frac{r}{2}\left\|c_{E}(x)\right\|_{2}^{2}+\mu_{I}^{\top}\left(c_{I}(x)+s\right)+\frac{r}{2}\left\|c_{I}(x)+s\right\|_{2}^{2}\right) .
$$

The augmented Lagrangian associated with $\left(P_{E I}\right)$ is the function of $x$ and $\mu$ defined by the minimal value of the optimization problem in $s \geq 0$ above:

$$
\begin{aligned}
\ell_{r}(x, \mu):=\min _{s \geq 0}(f(x) & +\mu_{E}^{\top} c_{E}(x)+\frac{r}{2}\left\|c_{E}(x)\right\|_{2}^{2} \\
& \left.+\mu_{I}^{\top}\left(c_{I}(x)+s\right)+\frac{r}{2}\left\|c_{I}(x)+s\right\|_{2}^{2}\right) .
\end{aligned}
$$

Actually, the minimization in $s$ can be carried out explicitly since the minimized function of $s$ is quadratic with a positive diagonal Hessian. More precisely, discarding terms independent of $s$, the objective can be written $\frac{r}{2}\left\|s+c_{I}(x)+\mu_{I} / r\right\|_{2}^{2}$, so that the minimizer is the projection of $-c_{I}(x)-\mu_{I} / r$ on the positive orthant, namely $s=\max \left(-c_{I}(x)-\mu_{I} / r, 0\right)$. Adding $c_{I}(x)$, one finds

$$
c_{I}(x)+s=\max \left(\frac{-\mu_{I}}{r}, c_{I}(x)\right) .
$$

Substituting $c_{I}(x)+s$ by this value in the objective of the problem above yields an explicit formula for the augmented Lagrangian. This is the function $\ell_{r}: \Omega \times \mathbb{R}^{m} \rightarrow \mathbb{R}$, defined for $(x, \mu) \in \Omega \times \mathbb{R}^{m}$ and $r \in \mathbb{R}_{+}^{*}:=\{t \in \mathbb{R}:$ $t>0\}$ by

$$
\begin{equation*}
\ell_{r}(x, \mu)=f(x)+\mu^{\top} \tilde{c}_{r}(x, \mu)+\frac{r}{2}\left\|\tilde{c}_{r}(x, \mu)\right\|_{2}^{2} \tag{16.6}
\end{equation*}
$$

where $\tilde{c}_{r}: \Omega \times \mathbb{R}^{m} \rightarrow \mathbb{R}^{m}$ is defined by

$$
\left(\tilde{c}_{r}(x, \mu)\right)_{i}= \begin{cases}c_{i}(x) & \text { if } i \in E  \tag{16.7}\\ \max \left(\frac{-\mu_{i}}{r}, c_{i}(x)\right) & \text { if } i \in I\end{cases}
$$

The coefficient $r$ is called the augmentation parameter. This augmented Lagrangian (16.6) has therefore a structure very similar to the one associated with the equality constraint problem $\left(P_{E}\right)$, see (16.5), with $c_{E}$ substituted by the non-differentiable function $\tilde{c}_{r}$ introduced above.

Despite the nonsmoothness of the max operator in (16.7), the augmented Lagrangian is differentiable in $x$, provided that $f$ and $c$ have that property. The easiest way of verifying this claim is to write the terms associated with the inequalities in (16.6) as follows

$$
\mu_{I}^{\top}\left(\tilde{c}_{r}(x, \mu)\right)_{I}+\frac{r}{2}\left\|\left(\tilde{c}_{r}(x, \mu)\right)_{I}\right\|_{2}^{2}=\frac{1}{2 r} \sum_{i \in I}\left(\max \left(0, \mu_{i}+r c_{i}(x)\right)^{2}-\mu_{i}^{2}\right)
$$

This is a differentiable function of $x$, since $\max (0, \cdot)$ is squared. A straightforward computation then leads to

$$
\begin{equation*}
\nabla_{x} \ell_{r}(x, \mu)=\nabla f(x)+c^{\prime}(x)^{\top}\left(\mu+r \tilde{c}_{r}(x, \mu)\right) \tag{16.8}
\end{equation*}
$$

Second differentiability in $x$ is also ensured around a primal solution satisfying some strong conditions. Let $x_{*}$ be a solution to $\left(P_{E I}\right)$ and let $\lambda_{*}$
be a multiplier associated with $x_{*}$. Using the complementarity conditions $\left(\lambda_{*}\right)_{I}^{\top} c_{I}\left(x_{*}\right)=0$ and the nonnegativity of $\left(\lambda_{*}\right)_{I}$, it is not difficult to see that, for $x$ close to $x_{*}$, there holds

$$
\begin{equation*}
\ell_{r}\left(x, \lambda_{*}\right)=\ell\left(x, \lambda_{*}\right)+\frac{r}{2} \sum_{i \in E \cup I_{*}^{0+}} c_{i}(x)^{2}+\frac{r}{2} \sum_{i \in I_{*}^{00}}\left(c_{i}(x)^{+}\right)^{2} \tag{16.9}
\end{equation*}
$$

Because of the operator $(\cdot)^{+}$in $(16.9), \ell_{r}\left(\cdot, \lambda_{*}\right)$ may not be twice differentiable at $x_{*}$. In the case of strict complementarity, however, $I_{*}^{00}=\emptyset$ and the last sum disappears, so that the augmented Lagrangian can be written (for $x$ close to $x_{*}$ )

$$
\ell_{r}\left(x, \lambda_{*}\right)=\ell\left(x, \lambda_{*}\right)+\frac{r}{2} \sum_{i \in E \cup I_{*}^{0}} c_{i}(x)^{2} .
$$

Locally, equality and active inequality constraints are then treated in the same way and $\ell_{r}\left(\cdot, \lambda_{*}\right)$ is smooth around $x_{*}$ (provided $f$ and $c$ are smooth). The next proposition gathers these differentiability properties.

Proposition 16.2 (differentiability of the augmented Lagrangian). If $f$ and $c$ are differentiable at $x$, then the augmented Lagrangian $\ell_{r}$, defined by (16.6), is differentiable at $x$ and its gradient is given by (16.8). If ( $x_{*}, \lambda_{*}$ ) is a KKT point for $\left(P_{E I}\right)$ satisfying strict complementarity and if $\left(f, c_{E \cup I_{*}^{0}}\right)$ is $p$ times differentiable (with $p \geq 0$ integer) in some neighborhood of $x_{*}$, then the augmented Lagrangian is $p$ times differentiable is some (possibly smaller) neighborhood of $x_{*}$.

The next result gives conditions for $\left(x_{*}, \lambda_{*}\right)$ to be a saddle-point of $\ell_{r}$ on $V \times \mathbb{R}^{m}$, where $V$ is a neighborhood of $x_{*}$ in $\Omega$. Compared with proposition 16.1, the result is local in $x$, but global in $\mu$, and the minimum in $x$ is strict. As before, this result implies that, if $r$ is large enough (but finite!), $\ell_{r}\left(\cdot, \lambda_{*}\right)$ is an exact penalty function for $\left(P_{E I}\right)$.

Proposition 16.3 (saddle-point of the augmented Lagrangian). Suppose that $f$ and $c_{E \cup I_{*}^{0}}$ are twice differentiable at a local minimum $x_{*}$ of $\left(P_{E I}\right)$ at which the KKT conditions hold, and that the semi-strong second-order sufficient condition of optimality (13.9) is satisfied for some multiplier $\lambda_{*}$. Then there exist a neighborhood $V$ of $x_{*}$ in $\Omega$ and a number $\underline{r}>0$ such that, for all $r \geq \underline{r},\left(x_{*}, \lambda_{*}\right)$ is a saddle-point of $\ell_{r}$ on $V \times \mathbb{R}^{m}$. More precisely, we have for all $(x, \mu) \in\left(V \backslash\left\{x_{*}\right\}\right) \times \mathbb{R}^{m}$ :

$$
\begin{equation*}
\ell_{r}\left(x_{*}, \mu\right) \leq \ell_{r}\left(x_{*}, \lambda_{*}\right)<\ell_{r}\left(x, \lambda_{*}\right) \tag{16.10}
\end{equation*}
$$

Proof. Let us first show that $\lambda_{*}$ maximizes $\ell_{r}\left(x_{*}, \cdot\right)$ for any $r>0$. We have for $\mu \in \mathbb{R}^{m}$ :

$$
\ell_{r}\left(x_{*}, \mu\right)=f\left(x_{*}\right)+\sum_{\substack{i \in I \\ c_{i}\left(x_{*}\right) \geq-\mu_{i} / r}}\left(\mu_{i} c_{i}\left(x_{*}\right)+\frac{r}{2} c_{i}\left(x_{*}\right)^{2}\right)-\sum_{\substack{i \in I \\ c_{i}\left(x_{*}\right)<-\mu_{i} / r}} \frac{\mu_{i}^{2}}{2 r}
$$

The maximum in $\mu$ can be obtained term by term. If $c_{i}\left(x_{*}\right)=0$, the maximum in the right-hand side is $f\left(x_{*}\right)$, obtained for all $\mu_{i} \geq 0$. If $c_{i}\left(x_{*}\right)<0$, this maximum is again $f\left(x_{*}\right)$, obtained for $\mu_{i}=0$. Since $\left(\lambda_{*}\right)_{I}$ satisfies these conditions, we have

$$
\ell_{r}\left(x_{*}, \mu\right) \leq f\left(x_{*}\right)=\ell_{r}\left(x_{*}, \lambda_{*}\right), \quad \text { for all } \mu \in \mathbb{R}^{m}
$$

Let us now show the second statement, dealing with the strict local minimality of $x_{*}$. Note that we need to prove the inequality on the right in (16.10) for only a single value of $r, \underline{r}>0$ say, because then, this inequality will hold for any $r \geq \underline{r}$ and any $x \in V$ (independent of $r$ ). Indeed, $\ell_{r}\left(x_{*}, \lambda_{*}\right)=f\left(x_{*}\right)$ does not depend on $r$ and, for fixed $x, r \mapsto \ell_{r}\left(x, \lambda_{*}\right)$ is nondecreasing (this is a clear consequence of the way the augmented Lagrangian was introduced, just before the proposition).

We prove this by contradiction, assuming that there is a sequence of positive numbers $r_{k} \rightarrow \infty$ and a sequence of points $x_{k} \rightarrow x_{*}$, with $x_{k} \neq x_{*}$ such that, for $k \geq 1$ :

$$
\begin{equation*}
\ell_{r_{k}}\left(x_{k}, \lambda_{*}\right) \leq \ell_{r_{k}}\left(x_{*}, \lambda_{*}\right) \tag{16.11}
\end{equation*}
$$

Taking a subsequence if necessary, we have for $k \rightarrow \infty$ :

$$
\frac{x_{k}-x_{*}}{\left\|x_{k}-x_{*}\right\|} \rightarrow d, \quad \text { with }\|d\|=1
$$

Hence, setting $\alpha_{k}:=\left\|x_{k}-x_{*}\right\|$, we have

$$
x_{k}=x_{*}+\alpha_{k} d+o\left(\alpha_{k}\right) .
$$

Our aim now is to show that $d$ is a critical direction. We do this by appropriate expansions in the left-hand side of (16.11): second order expansion of the Lagrangian and first order expansion of the constraints in both sums of (16.9). To simplify the notation, we introduce $L_{*}=\nabla_{x x}^{2} \ell\left(x_{*}, \lambda_{*}\right)$. From the smoothness of $f$ and $c$ and the optimality of $\left(x_{*}, \lambda_{*}\right)$, we have

$$
\begin{aligned}
\ell\left(x_{k}, \lambda_{*}\right) & =\ell\left(x_{*} \lambda_{*}\right)+\frac{\alpha_{k}^{2}}{2} d^{\top} L_{*} d+o\left(\alpha_{k}^{2}\right) \\
c_{i}\left(x_{k}\right) & =\alpha_{k} c_{i}^{\prime}\left(x_{*}\right) \cdot d+o\left(\alpha_{k}\right), \quad \text { for } i \in E \cup I_{*}^{0} .
\end{aligned}
$$

Injecting these estimates in (16.11), using (16.9) and $\ell_{r_{k}}\left(x_{*}, \lambda_{*}\right)=\ell\left(x_{*}, \lambda_{*}\right)$, provides

$$
\begin{align*}
\frac{\alpha_{k}^{2}}{2} d^{\top} L_{*} d+o\left(\alpha_{k}^{2}\right) & +\frac{r_{k}}{2} \sum_{i \in E \cup I_{*}^{0+}}\left(\alpha_{k} c_{i}^{\prime}\left(x_{*}\right) \cdot d+o\left(\alpha_{k}\right)\right)^{2} \\
& +\frac{r_{k}}{2} \sum_{i \in I_{*}^{00}}\left(\left[\alpha_{k} c_{i}^{\prime}\left(x_{*}\right) \cdot d+o\left(\alpha_{k}\right)\right]^{+}\right)^{2} \leq 0 \tag{16.12}
\end{align*}
$$

Dividing by $\alpha_{k}^{2} r_{k}$ and taking the limit yield

$$
\begin{array}{ll}
c_{i}^{\prime}\left(x_{*}\right) \cdot d=0, & \text { if } i \in E \cup I_{*}^{0+} \\
c_{i}^{\prime}\left(x_{*}\right) \cdot d \leq 0, & \text { if } i \in I_{*}^{00} .
\end{array}
$$

Therefore $d$ is a nonzero critical direction.
On the other hand, (16.12) also implies that

$$
\frac{\alpha_{k}^{2}}{2} d^{\top} L_{*} d+o\left(\alpha_{k}^{2}\right) \leq 0
$$

Dividing by $\alpha_{k}^{2}$ and taking the limit show that $d^{\top} L_{*} d \leq 0$, which contradicts assumption (13.9) since $d \in C_{*} \backslash\{0\}$.

In the previous result, the semi-strong second-order sufficient condition of optimality (13.9) is assumed. If only the weak condition (13.8) holds, $\ell_{r}\left(\cdot, \lambda_{*}\right)$ may not have a local minimum at $x_{*}$, whatever the choice of $\lambda_{*} \in \Lambda_{*}$ and the value of $r$. An example is given in exercise 16.4.

### 16.4 Nondifferentiable Augmented Function

We now consider the following merit function for problem $\left(P_{E I}\right)$ :

$$
\begin{equation*}
\Theta_{\sigma}(x)=f(x)+\sigma\left\|c(x)^{\#}\right\|_{P}, \tag{16.13}
\end{equation*}
$$

which we call the nondifferentiable augmented function. In (16.13), $\sigma>0$ is called the penalty parameter, the operator $(\cdot)^{\#}$ was defined on page 194, and $\|\cdot\|_{P}$ is a norm, and is arbitrary for the moment. We denote by $\|\cdot\|_{D}$ the dual norm of $\|\cdot\|_{P}$, with respect to the Euclidean scalar product. It is defined by

$$
\|v\|_{D}=\sup _{\|u\|_{P}=1} v^{\top} u
$$

We therefore have the generalized Cauchy-Schwarz inequality:

$$
\begin{equation*}
\left|u^{\top} v\right| \leq\|u\|_{P}\|v\|_{D}, \quad \text { for all } u \text { and } v \tag{16.14}
\end{equation*}
$$

See exercise 16.5 for some examples of dual norms.
Because of the norm $\|\cdot\|_{P}$ and of the operator $(\cdot)^{\#}, \Theta_{\sigma}$ is usually nondifferentiable; but when $f$ and $c$ are smooth, $\Theta_{\sigma}$ has directional derivatives; this is a consequence of lemma 13.1. It so happens that this differentiability concept will be sufficient for our development.

Let $v \in \mathbb{R}^{m}$ be such that $v_{I} \leq 0$ and denote by $P_{v}: \mathbb{R}^{m} \rightarrow \mathbb{R}^{m}$ the operator defined by $P_{v} u=\left(.{ }^{\#}\right)^{\prime}(v ; u)$, that is

$$
\left(P_{v} u\right)_{i}=\left\{\begin{array}{l}
u_{i} \text { si } i \in E \\
u_{i}^{+} \text {if } i \in I \text { and } v_{i}=0 \\
0 \quad \text { if } i \in I \text { and } v_{i}<0
\end{array}\right.
$$

This notation allows us to write concisely the directional derivative of $\Theta_{\sigma}$ at a feasible point.

Lemma 16.4. If $f$ and $c$ have a directional derivative at $x$ in the direction $h \in \mathbb{R}^{n}$, then $\Theta_{\sigma}$ has also a directional derivative at $x$ in the direction $h$. If, in addition, $x$ is feasible for $\left(P_{E I}\right)$, we have

$$
\Theta_{\sigma}^{\prime}(x ; h)=f^{\prime}(x ; h)+\sigma\left\|P_{c(x)} c^{\prime}(x ; h)\right\|_{P}
$$

Proof. The directional differentiability of $\Theta_{\sigma}=f+\sigma\left(\|\cdot\|_{P} \circ(\cdot)^{\#} \circ c\right)$ comes from lemma 13.1, the assumptions on $f$ and $c$, and the fact that $(\cdot)^{\#}$ and $\|\cdot\|_{P}$ are Lipschitzian and have directional derivatives.

If $x$ is feasible, $c(x)^{\#}=0$ and we have from lemma 13.1,

$$
\Theta_{\sigma}^{\prime}(x ; h)=f^{\prime}(x ; h)+\sigma\left(\|\cdot\|_{P}\right)^{\prime}\left(0 ;\left(c^{\#}\right)^{\prime}(x ; h)\right) .
$$

On the other hand,

$$
\left(c^{\#}\right)^{\prime}(x ; h)=(. \#)^{\prime}\left(c(x) ; c^{\prime}(x ; h)\right)=P_{c(x)} c^{\prime}(x ; h)
$$

and

$$
\left(\|\cdot\|_{P}\right)^{\prime}(0 ; v)=\lim _{t \rightarrow 0+} \frac{1}{t}\left(\|t v\|_{P}-0\right)=\|v\|_{P} .
$$

The result follows.

## Necessary Conditions of Exactness

In this subsection, we examine which conditions are implied by the fact that a feasible point $x_{*}$ is a minimum point of $\Theta_{\sigma}$. We quote three such properties in proposition 16.5: $x_{*}$ is also a minimum point of $\left(P_{E I}\right)$, there exists a multiplier $\lambda_{*}$ associated with $x_{*}$, and $\sigma$ must be sufficiently large. The second property shows that the exactness of $\Theta_{\sigma}$ plays a similar role as a constraint qualification assumption, since it implies the existence of a dual solution.

For the third property mentioned above, we need an assumption on the norm $\|\cdot\|_{P}$ used in $\Theta_{\sigma}$. The norm $\|v\|_{P}$ must decrease if one sets to zero some of the $I$-components of $v \in \mathbb{R}^{m}$ :

$$
u_{i}=\left\{\begin{array}{ll}
v_{i} & \text { if } i \in E  \tag{16.15}\\
0 \text { or } v_{i} & \text { if } i \in I
\end{array} \quad \Longrightarrow \quad\|u\|_{P} \leq\|v\|_{P}\right.
$$

Clearly, $\ell_{p}$ norms, $1 \leq p \leq \infty$, satisfy this property; but it is not necessarily satisfied by an arbitrary norm (see exercise 17.1). Also, the claim on $\sigma$ in proposition 16.5 may not be correct if $\|\cdot\|_{P}$ does not satisfy (16.15) (see exercise 16.6).

Proposition 16.5 (necessary conditions of exactness). If $x_{*}$ is feasible for $\left(P_{E I}\right)$ and $\Theta_{\sigma}$ has a local minimum (resp. strict local minimum) at $x_{*}$, then $x_{*}$ is a local minimum (resp. strict local minimum) of $\left(P_{E I}\right)$. If, in addition, $f$ and $c$ are Gâteaux differentiable at $x_{*}$, then there exists a multiplier $\lambda_{*}$ such that the necessary optimality conditions (KKT) hold. If, in addition, the norm $\|\cdot\|_{P}$ satisfies (16.15) and (LI-CQ) holds at $x_{*}$, then $\sigma \geq\left\|\lambda_{*}\right\|_{D}$.

Proof. If $x_{*}$ is a local minimum of $\Theta_{\sigma}$, there exists a neighborhood $V$ of $x_{*}$ such that

$$
\Theta_{\sigma}\left(x_{*}\right) \leq \Theta_{\sigma}(x), \quad \text { for all } x \in V .
$$

Since $x_{*} \in X$ and $\left.\Theta_{\sigma}\right|_{X}=\left.f\right|_{X}$, we have

$$
f\left(x_{*}\right) \leq f(x), \quad \text { for all } x \in V \cap X,
$$

which shows that $x_{*}$ is a local minimum of $\left(P_{E I}\right)$. The above inequality is strict for $x \neq x_{*}$, if $x_{*}$ is a strict local minimum of $\Theta_{\sigma}$.

Now suppose $f$ and $g$ are Gâteaux differentiable at $x_{*}$. Then $\Theta_{\sigma}$ has directional derivatives at $x_{*}$ (lemma 16.4). Since $x_{*}$ is a local minimum of $\Theta_{\sigma}$, we have $\Theta_{\sigma}^{\prime}\left(x_{*} ; h\right) \geq 0$ for all $h \in \mathbb{R}^{m}$. But $x_{*}$ is feasible; hence, by lemma 16.4:

$$
\begin{equation*}
\nabla f\left(x_{*}\right)^{\top} h+\sigma\left\|P_{c\left(x_{*}\right)}\left(A\left(x_{*}\right) h\right)\right\|_{P} \geq 0, \quad \text { for all } h \in \mathbb{R}^{m} \tag{16.16}
\end{equation*}
$$

We deduce

$$
P_{c\left(x_{*}\right)}\left(A\left(x_{*}\right) h\right)=0 \quad \Longrightarrow \quad \nabla f\left(x_{*}\right)^{\top} h \geq 0
$$

Thus, $h=0$ solves the linear program

$$
\left\{\begin{array}{l}
\min _{h} \nabla f\left(x_{*}\right)^{\top} h \\
A_{E}\left(x_{*}\right) h=0 \\
A_{I_{*}^{0}}\left(x_{*}\right) h \leq 0
\end{array}\right.
$$

The constraints of this problem being qualified (by (A-CQ)), we deduce the existence of a multiplier $\lambda_{*} \in \mathbb{R}^{m}$ such that

$$
\left\{\begin{array}{l}
\nabla f\left(x_{*}\right)+A\left(x_{*}\right)^{\top} \lambda_{*}=0 \\
\left(\lambda_{*}\right)_{I_{*}^{0}} \geq 0 \\
\left(\lambda_{*}\right)_{I \backslash I_{*}^{0}}=0
\end{array}\right.
$$

Since $x_{*}$ is feasible, (KKT) holds with $\left(x_{*}, \lambda_{*}\right)$.
Finally, suppose that the norm $\|\cdot\|_{P}$ satisfies (16.15) and that (LI-CQ) holds. Take again (16.16) and use the first-order optimality condition to obtain

$$
\lambda_{*}^{\top} A\left(x_{*}\right) h \leq \sigma\left\|P_{c\left(x_{*}\right)} A\left(x_{*}\right) h\right\|_{P}, \quad \text { for all } h \in \mathbb{R}^{n}
$$

Set $J=E \cup I_{*}^{0}$, and remember that $\left(\lambda_{*}\right)_{i}=0$ if $i \notin J$. For an arbitrary $v$ in $\mathbb{R}^{m}$, we have $\lambda_{*}^{\top} v=\left(\lambda_{*}\right)_{J}^{\top} v_{J}$ and, from (LI-CQ), we can find $h \in \mathbb{R}^{n}$ such that $A_{J}\left(x_{*}\right) h=v_{J}$. We deduce that

$$
\lambda_{*}^{\top} v=\left(\lambda_{*}\right)_{J}^{\top} A_{J}\left(x_{*}\right) h=\lambda_{*}^{\top} A\left(x_{*}\right) h \leq \sigma\left\|P_{c\left(x_{*}\right)} A\left(x_{*}\right) h\right\|_{P} \leq \sigma\|v\|_{P},
$$

where the last inequality uses property (16.15) of the norm. Then $\lambda_{*}^{\top} v \leq$ $\sigma\|v\|_{P}$, and since $v$ is arbitrary, we have $\left\|\lambda_{*}\right\|_{D} \leq \sigma$.

## Sufficient Conditions of Exactness

In practice, we are more interested in having conditions that ensure the exactness of $\Theta_{\sigma}$ and this is what we focus on now. We shall show that the necessary condition obtained on $\sigma$ in proposition 16.5 is sharp: if $x_{*}$ is a strong solution to problem $\left(P_{E I}\right)$ with associated multiplier $\lambda_{*}, x_{*}$ also minimizes $\Theta_{\sigma}$ provided $\sigma>\left\|\lambda_{*}\right\|_{D}$ (the strict inequality is not needed for convex problems). This result holds without any particular assumption on the norm $\|\cdot\|_{P}$.

The necessary conditions of exactness of $\Theta_{\sigma}$ were obtained by expressing the fact that, if $x_{*}$ minimizes $\Theta_{\sigma}$, the directional derivative $\Theta_{\sigma}^{\prime}\left(x_{*} ; h\right)$ must be nonnegative for all $h \in \mathbb{R}^{n}$ (see the proof of proposition 16.5). Now we want to exhibit values of $\sigma$ such that $\Theta_{\sigma}$ has a minimum at $x_{*}$. Function $\Theta_{\sigma}$ is nondifferentiable and nonconvex. Therefore, it is not sufficient to show that $\Theta_{\sigma}^{\prime}\left(x_{*} ; h\right) \geq 0$ for all $h \in \mathbb{R}^{n}$ in order to ensure its exactness. One cannot impose $\Theta_{\sigma}^{\prime}\left(x_{*} ; h\right)>0$ for all $h \in \mathbb{R}^{n}$ either, since this may never occur for any value of $\sigma$ (for example, when $E \neq \emptyset$ and $I=\emptyset, \Theta_{\sigma}^{\prime}\left(x_{*} ; h\right)=0$ for any $h$ in the space tangent to the constraint manifold). Therefore, we shall use either a technical detour (for convex problems) or a direct proof like the one of proposition 16.3 (for nonconvex problems).

In proposition 16.7 below, we consider the case of convex problems and in proposition 16.8 the case of nonconvex problems. To prove the exactness of the nondifferentiable function $\Theta_{\sigma}$ for convex problems, we simply use the fact that, if $\sigma$ is large enough, $\Theta_{\sigma}$ is above the differentiable Lagrangian (16.4) (lemma 16.6), which is known to be exact at $x_{*}$ (proposition 16.1). Observe that lemma 16.6 does not assume convexity.
Lemma 16.6. If $\sigma \geq\|\lambda\|_{D}$ and $\lambda_{I} \geq 0$, then $\ell(\cdot, \lambda) \leq \Theta_{\sigma}(\cdot)$ on $\mathbb{R}^{n}$.
Proof. First observe that $\lambda_{I} \geq 0$ implies $\lambda_{I}^{\top} c_{I}(x) \leq \lambda_{I}^{\top} c_{I}(x)^{+}$. Then, for all $x \in \mathbb{R}^{n}$,

$$
\ell(x, \lambda) \leq f(x)+\lambda^{\top} c(x)^{\#} \leq f(x)+\|\lambda\|_{D}\left\|c(x)^{\#}\right\|_{P} \leq \Theta_{\sigma}(x)
$$

Proposition 16.7 (sufficient conditions of exactness, convex problems). Suppose that problem $\left(P_{E I}\right)$ is convex and that $f$ and $c$ are differentiable at a solution $x_{*}$ to $\left(P_{E I}\right)$ with an associated multiplier $\lambda_{*}$. Then $\Theta_{\sigma}$ has a global minimum at $x_{*}$ as soon as $\sigma \geq\left\|\lambda_{*}\right\|_{D}$.

Proof. According to proposition 16.1, $\ell\left(\cdot, \lambda_{*}\right)$ is minimized by $x_{*}$ and, by lemma 16.6, it is dominated by $\Theta_{\sigma}\left(\sigma \geq\left\|\lambda_{*}\right\|_{D}\right.$ and $\left.\left(\lambda_{*}\right)_{I} \geq 0\right)$. Therefore

$$
\begin{aligned}
\Theta_{\sigma}\left(x_{*}\right) & =f\left(x_{*}\right) \\
& =\ell\left(x_{*}, \lambda_{*}\right) \\
& \leq \ell\left(x, \lambda_{*}\right), \quad \text { for all } x \in \mathbb{R}^{n} \\
& \leq \Theta_{\sigma}(x), \quad \text { for all } x \in \mathbb{R}^{n}
\end{aligned}
$$

The same technical detour could be used for highlighting sufficient conditions of exactness of $\Theta_{\sigma}$ for nonconvex problems: if $\sigma>\left\|\lambda_{*}\right\|_{D}, \Theta_{\sigma}$ is above the augmented Lagrangian (16.6) in some neighborhood of $x_{*}$, so that the exactness of $\Theta_{\sigma}$ follows that of the augmented Lagrangian (proposition 16.3). This strategy is proposed in exercise 16.7. The direct proof given below has the advantage of being valid even when only the weak second order sufficient condition of optimality (13.8) holds at $x_{*}$ (in contrast, the semi-strong condition (13.9) is assumed in proposition 16.3 and exercise 16.7).

Proposition 16.8 (sufficient conditions of exactness). Suppose that $f$ and $c_{E \cup I_{*}^{0}}$ are twice differentiable at a local minimum $x_{*}$ of $\left(P_{E I}\right)$ at which the KKT conditions hold, that the weak second-order sufficient condition of optimality (13.8) is satisfied, and that

$$
\sigma>\sup _{\lambda_{*} \in \Lambda_{*}}\left\|\lambda_{*}\right\|_{D}
$$

where $\Lambda_{*}$ is the nonempty set of multipliers associated with $x_{*}$. Then $\Theta_{\sigma}$ has a strict local minimum at $x_{*}$.

Proof. We prove the result by contradiction, assuming that $x_{*}$ is not a strict minimum of $\Theta_{\sigma}$. Then, there exists a sequence $\left\{x_{k}\right\}$ such that $x_{k} \neq x_{*}$, $x_{k} \rightarrow x_{*}$ and

$$
\begin{equation*}
\Theta_{\sigma}\left(x_{k}\right) \leq \Theta_{\sigma}\left(x_{*}\right), \quad \forall k \geq 1 . \tag{16.17}
\end{equation*}
$$

Since the sequence $\left\{\left(x_{k}-x_{*}\right) /\left\|x_{k}-x_{*}\right\|\right\}$ is bounded (here $\|\cdot\|$ denotes an arbitrary norm), it has a subsequence such that $\left(x_{k}-x_{*}\right) /\left\|x_{k}-x_{*}\right\| \rightarrow d$, where $\|d\|=1$. Denoting $\alpha_{k}=\left\|x_{k}-x_{*}\right\|$, one has

$$
x_{k}=x_{*}+\alpha_{k} d+o\left(\alpha_{k}\right)
$$

Because $\Theta_{\sigma}$ is Lipschitzian in a neighborhood of $x_{*}$ :

$$
\Theta_{\sigma}\left(x_{k}\right)=\Theta_{\sigma}\left(x_{*}+\alpha_{k} d\right)+o\left(\alpha_{k}\right) .
$$

Now (16.17) shows that $\Theta_{\sigma}^{\prime}\left(x_{*} ; d\right) \leq 0$. Then, from lemma 16.4 , one can write

$$
\begin{equation*}
f^{\prime}\left(x_{*}\right) \cdot d+\sigma\left\|P_{c\left(x_{*}\right)}\left(c^{\prime}\left(x_{*}\right) \cdot d\right)\right\|_{P} \leq 0 . \tag{16.18}
\end{equation*}
$$

This certainly implies that

$$
\begin{equation*}
f^{\prime}\left(x_{*}\right) \cdot d \leq 0 . \tag{16.19}
\end{equation*}
$$

On the other hand, from the assumptions, there is an optimal multiplier $\lambda_{*}$ such that $\sigma>\left\|\lambda_{*}\right\|_{D}$. Using the first order optimality conditions, including the nonnegativity of $\left(\lambda_{*}\right)_{I}$ and the complementarity conditions $\left(\lambda_{*}\right)_{I}^{\top} c_{I}\left(x_{*}\right)=0$, one has

$$
\begin{aligned}
-f^{\prime}\left(x_{*}\right) \cdot d & =\lambda_{*}^{\top}\left(c^{\prime}\left(x_{*}\right) \cdot d\right) \\
& \leq \lambda_{*}^{\top} P_{c\left(x_{*}\right)}\left(c^{\prime}\left(x_{*}\right) \cdot d\right) \\
& \leq\left\|\lambda_{*}\right\|_{D}\left\|P_{c\left(x_{*}\right)}\left(c^{\prime}\left(x_{*}\right) \cdot d\right)\right\|_{P} .
\end{aligned}
$$

Then (16.18) and $\sigma>\left\|\lambda_{*}\right\|_{D}$ imply that $P_{c\left(x_{*}\right)}\left(c^{\prime}\left(x_{*}\right) \cdot d\right)=0$, i.e.,

$$
\left\{\begin{array}{l}
c_{i}^{\prime}\left(x_{*}\right) \cdot d=0 \text { for } i \in E \\
c_{i}^{\prime}\left(x_{*}\right) \cdot d \leq 0 \text { for } i \in I_{*}^{0} .
\end{array}\right.
$$

These and (16.19) show that $d$ is a nonzero critical direction.
Now, let $\lambda_{*}$ be the multiplier depending on $d$, determined by the weak second-order sufficient condition of optimality (13.8). According to theorem 13.4, one has

$$
d^{\top} \nabla_{x x}^{2} \ell\left(x_{*}, \lambda_{*}\right) d>0
$$

The following Taylor expansion (use $\nabla_{x} \ell\left(x_{*}, \lambda_{*}\right)=0$ )

$$
\ell\left(x_{k}, \lambda_{*}\right)=\ell\left(x_{*}, \lambda_{*}\right)+\frac{\alpha_{k}^{2}}{2} d^{\top} \nabla_{x x}^{2} \ell\left(x_{*}, \lambda_{*}\right) d+o\left(\alpha_{k}^{2}\right)
$$

allows us to see that, for $k$ large enough,

$$
\begin{equation*}
\ell\left(x_{k}, \lambda_{*}\right)>\ell\left(x_{*}, \lambda_{*}\right) \tag{16.20}
\end{equation*}
$$

Then, for large indices $k$, there holds

$$
\begin{aligned}
\Theta_{\sigma}\left(x_{k}\right) & \leq \Theta_{\sigma}\left(x_{*}\right) \quad[\text { by }(16.17)] \\
& =f\left(x_{*}\right) \\
& =\ell\left(x_{*}, \lambda_{*}\right) \\
& <\ell\left(x_{k}, \lambda_{*}\right) \quad[\text { by }(16.20)] \\
& \leq \Theta_{\sigma}\left(x_{k}\right) \quad\left[\text { by lemma } 16.6 \text { and } \sigma \geq\left\|\lambda_{*}\right\|_{D}\right],
\end{aligned}
$$

which is the expected contradiction.

## Notes

The augmented Lagrangian (16.2) for equality constrained problems was first proposed by Arrow and Solow [14; 1958]. Hestenes [191; 1969] and Powell [288; 1969] both used this function to introduce the so-called method of multipliers, which has popularized this type of penalization. The augmented Lagrangian (16.3) or (16.6), adapted to inequality constrained problems, was proposed by Rockafellar [310, 311; 1971-74] and Buys [62; 1972]. It was further extended to constraints of the form $c(x) \in K$, where $c$ is a vector-valued function and
$K$ is a closed convex cone, by Shapiro and Sun [330; 2004]. This penalty function is usually no more than continuously differentiable, even if the problem data are infinitely differentiable. Many developments have been carried out to overcome this drawback, proposing augmentation terms with a different structure (for entry points see [17, 16; 1999-2000], which deal with primal penalty functions, and [109, 110, 111; 1999-2001], which consider primal-dual penalty functions). Surveys on the augmented Lagrangian can be found in [26, 169].

The exact penalty function (16.13) goes back at least to Eremin [119; 1966] and Zangwill $[374 ; 1967]$. Its connection with problem $\left(P_{E I}\right)$ has been studied by many authors, see Pietrzykowski [284], Charalambous [74], Ioffe [198], Han and Mangasarian [186], Bertsekas [26], Fletcher [126], Bonnans [39, 41], Facchinei [120], Burke [60], Pshenichnyj [301], Bonnans and Shapiro [50], and the references therein.

## Exercises

16.1. Finsler's lemma [123] and its limit case [9]. Let $M$ be an $n \times n$ symmetric matrix that is positive definite on the null space of a matrix $A$ (i.e., $u^{\top} M u>0$ for all nonzero $\left.u \in N(A)\right)$. Show that there exists an $r_{0} \in \mathbb{R}$ such that, for all $r \geq r_{0}, M+r A^{\top} A$ is positive definite.
[Hint: Use an argument by contradiction.]
Suppose now that the symmetric matrix $M$ is only positive semidefinite on the null space of $A$ (i.e., $u^{\top} M u \geq 0$ for all $u \in N(A)$ ). Show that the following claims are equivalent: $(i) v \in N(A)$ and $v^{\top} M v=0$ imply that $M v=0$, and (ii) there exists an $r_{0} \in \mathbb{R}$ such that, for all $r \geq r_{0}, M+r A^{\top} A$ is positive semidefinite. Find a matrix $M$ that is positive semidefinite on the null space of $A$, for which these properties $(i)$ and $(i i)$ are not satisfied.
[Hint: For $(i) \Rightarrow(i i)$, use with care an argument by contradiction.]
Consequence: If $M$ is nonsingular and positive semidefinite (but not positive definite) on the null space of $A$, it cannot enjoy property (ii) (since ( $i$ ) does not hold).
16.2. Augmented Lagrangian for equality constrained problems. Consider problem $\left(P_{E}\right)$ with functions $f$ and $c$ of class $C^{2}$ and the associated augmented Lagrangian $\ell_{r}(x, \lambda)=f(x)+\lambda^{\top} c(x)+\frac{r}{2}\|c(x)\|_{2}^{2}$. By a direct computation of $\nabla_{x} \ell_{r}\left(x_{*}, \lambda_{*}\right)$ and $\nabla_{x x}^{2} \ell_{r}\left(x_{*}, \lambda_{*}\right)$, show that, if $r$ is large enough, $\ell_{r}\left(\cdot, \lambda_{*}\right)$ has a strict local minimum at a point $x_{*}$ satisfying (SC2).
[Hint: Use Finsler's lemma (exercise 16.1).]
16.3. Fletcher's exact penalty function [124]. Consider problem $\left(P_{E}\right)$, in which $f$ and $c$ are smooth, and $c$ is a submersion. Denote by $A^{-}(x)$ a right inverse of the constraint Jacobian $A(x):=c^{\prime}(x)$ and assume that $A^{-}$is a smooth function of $x$. Let $\lambda^{\text {LS }}(x):=-A^{-}(x)^{\top} \nabla f(x)$ be the associated least-squares multiplier. For $r \in \mathbb{R}$, consider the function $\varphi_{r}: \mathbb{R}^{n} \rightarrow \mathbb{R}$ defined by

$$
\begin{equation*}
\varphi_{r}(x)=f(x)+\lambda^{\mathrm{LS}}(x)^{\top} c(x)+\frac{r}{2}\|c(x)\|_{2}^{2} . \tag{16.21}
\end{equation*}
$$

Let $\left(x_{*}, \lambda_{*}\right)$ be a pair satisfying the second-order sufficient conditions of optimality (SC2) of problem $\left(P_{E}\right)$. Show that there exists an $r_{0} \in \mathbb{R}$, such that for $r \geq r_{0}, \varphi_{r}$ has a strict local minimum at $x_{*}$.
[Hint: Prove the following claims, in which $A_{*}:=A\left(x_{*}\right), A_{*}^{-}:=A^{-}\left(x_{*}\right)$, and $L_{*}:=\nabla_{x x}^{2} \ell\left(x_{*}, \lambda_{*}\right)$, and conclude: $(i) \lambda^{\mathrm{LS}}\left(x_{*}\right)=\lambda_{*} ;(i i) \nabla \varphi_{r}\left(x_{*}\right)=0$; (iii) $\left(\lambda^{\mathrm{LS}}\right)^{\prime}\left(x_{*}\right)=-A_{*}^{-\top} L_{*}$ and $\nabla^{2} \varphi_{r}\left(x_{*}\right)=L_{*}-\left(A_{*}^{\top} A_{*}^{-\top} L_{*}+L_{*} A_{*}^{-} A_{*}\right)+$ $r A_{*}^{\top} A_{*} ;(i v) \nabla^{2} \varphi_{r}\left(x_{*}\right)$ is positive definite if $r$ is large enough.]
16.4. Counter-example for proposition 16.3 . Consider the problem in $\mathbb{R}^{3}$ :

$$
\left\{\begin{array}{l}
\min _{x} x_{3} \\
x_{3} \geq\left(x_{1}+x_{2}\right)\left(x_{1}-x_{2}\right) \\
x_{3} \geq\left(x_{2}+3 x_{1}\right)\left(2 x_{2}-x_{1}\right) \\
x_{3} \geq\left(2 x_{2}+x_{1}\right)\left(x_{2}-3 x_{1}\right)
\end{array}\right.
$$

Show that: $(i) x_{*}=0$ is the unique solution to the problem and that the associated multiplier set is $\Lambda_{*}=\left\{\lambda \in \mathbb{R}_{+}^{3}: \lambda_{1}+\lambda_{2}+\lambda_{3}=1\right\} ;(i i)$ the weak second order sufficient condition of optimality (13.8) is satisfied, but not the semi-strong ones (13.9); (iii) for any $\lambda_{*} \in \Lambda_{*}$ and $r \geq 0$, the augmented Lagrangian (16.6) has not a minimum at $x_{*}$.
Consequence: When the semi-strong second order sufficient conditions of optimality (13.9) do not hold at $x_{*}$, the augmented Lagrangian $\ell_{r}\left(\cdot, \lambda_{*}\right)$ function may not have a local minimum at $x_{*}$, for any $\lambda_{*} \in \Lambda_{*}$ and $r \geq 0$.
16.5. Dual norms. (i) The $\ell_{p}$ norm on $\mathbb{R}^{n}$ is defined by

$$
\|u\|_{p}:= \begin{cases}\left(\sum_{i=1}^{n}\left|u_{i}\right|^{p}\right)^{\frac{1}{p}} & \text { if } 1 \leq p<\infty \\ \max _{1 \leq i \leq n}\left|u_{i}\right| & \text { if } p=\infty .\end{cases}
$$

Show that the dual norm of $\|\cdot\|_{p}$ is the norm $\|\cdot\|_{p^{\prime}}$, where $p^{\prime}$ is uniquely defined by

$$
\frac{1}{p}+\frac{1}{p^{\prime}}=1
$$

(ii) Let $Q$ be a symmetric positive definite matrix and define the norm $\|u\|_{P}=\left(u^{\top} Q u\right)^{\frac{1}{2}}$. Show that its dual norm is given by $\|v\|_{D}=\left(v^{\top} Q^{-1} v\right)^{\frac{1}{2}}$.
16.6. Counter-example for proposition 16.5. Consider the problem

$$
\min \left\{\frac{1}{2}\|x\|_{2}^{2}: x \in \mathbb{R}^{2}, x_{1} \leq 0, x_{2}+1 \leq 0\right\} .
$$

Show that the unique primal-dual solution to this problem is $x_{*}=(0,-1)$ and $\lambda_{*}=(0,1)$. Show that $x \mapsto\|x\|_{P}=\left(x_{1}^{2}+x_{2}^{2}+\sqrt{3} x_{1} x_{2}\right)^{1 / 2}$ is a norm that does not satisfy (16.15), and that $\left\|\lambda_{*}\right\|_{D}=2$. Show that $\Theta_{\sigma}(x)=$ $\frac{1}{2}\|x\|_{2}^{2}+\sigma\left\|\left(x_{1}, x_{2}+1\right)^{+}\right\|_{P}$ has a minimum at $x_{*}$ when $\sigma \geq 1$.
Consequence: The exactness of $\Theta_{\sigma}$ does not imply $\sigma \geq\left\|\lambda_{*}\right\|_{D}$ if the norm $\|\cdot\|_{P}$ does not satisfy (16.15).
16.7. A variant of proposition 16.8. (i) Let $x_{*}$ be feasible for $\left(P_{E I}\right)$ and $\lambda \in \mathbb{R}^{m}$ be such that $\lambda_{I} \geq 0$ and $\lambda_{I}^{\top} c_{I}\left(x_{*}\right)=0$; let $r>0$ and $\sigma>\|\lambda\|_{D}$. Show that there exists a neighborhood $V$ of $x_{*}$ in $\Omega$ such that for all $x \in V$, there holds $\ell_{r}(x, \lambda) \leq \Theta_{\sigma}(x)$.
(ii) Suppose that $f$ and $c_{E \cup I_{*}^{0}}$ are twice differentiable at a local minimum $x_{*}$ of ( $P_{E I}$ ) at which the $\mathrm{KKT}^{*}$ conditions hold, that the semi-strong secondorder sufficient condition of optimality (13.9) holds for some optimal multiplier $\lambda_{*}$, and that $\sigma>\left\|\lambda_{*}\right\|_{D}$. Show, using $(i)$, that $\Theta_{\sigma}$ has a strict local minimum at $x_{*}$.
16.8. $\ell_{1}$ penalty function. Suppose that $f$ and $c_{E \cup I_{*}^{0}}$ are twice differentiable at a local minimum $x_{*}$ of $\left(P_{E I}\right)$ at which the KKT conditions hold and that the weak second-order sufficient condition of optimality (13.8) is satisfied. Positive scalars $\sigma_{i}(i \in E \cup I)$ are given and the following penalty function is considered:

$$
\Theta_{\sigma}^{1}(x)=f(x)+\sum_{i \in E} \sigma_{i}\left|c_{i}(x)\right|+\sum_{i \in I} \sigma_{i} c_{i}(x)^{+} .
$$

Show that, if $\sigma_{i}>\left|\left(\lambda_{*}\right)_{i}\right|$, for $i \in E \cup I$ and all optimal multiplier $\lambda_{*}$, then $x_{*}$ is a strict local minimum of $\Theta_{\sigma}^{1}$.
[Hint: Use the norm $v \mapsto\|v\|_{P}:=\sum_{i} \sigma_{i}\left|v_{i}\right|$ and proposition 16.8.]
Remark: The $\ell_{1}$-penalty function offers a natural way of controlling the magnitude of penalty parameters $\sigma_{i}$, when one such parameter is associated with each constraint.
16.9. Nondifferentiable augmented Lagrangian ([37] for equality constrained problems; [41] for an alternative to (16.22)). Suppose that $f$ and $c_{E \cup I_{*}^{0}}$ are twice differentiable at a local minimum $x_{*}$ of $\left(P_{E I}\right)$ at which the KKT conditions hold. Let be given $\mu \in \mathbb{R}^{m}$ and $\sigma \in \mathbb{R}_{+}$. Suppose one of the following:
(i) either the weak second-order sufficient condition of optimality (13.8) is satisfied and $\sigma>\sup \left\{\left\|\lambda_{*}-\mu\right\|_{D}: \lambda_{*} \in \Lambda_{*}\right\}$,
(ii) or the semi-strong second-order sufficient condition of optimality (13.9) holds for some optimal multiplier $\lambda_{*}$ and $\sigma>\left\|\mu-\lambda_{*}\right\|_{D}$.

Then $\Theta_{\mu, \sigma}: \mathbb{R}^{n} \rightarrow \mathbb{R}$ defined by

$$
\begin{equation*}
\Theta_{\mu, \sigma}(x):=f(x)+\mu^{\top} c(x)^{\#}+\sigma\left\|c(x)^{\#}\right\|_{P} \tag{16.22}
\end{equation*}
$$

has a strict local minimum at $x_{*}$.
[Hint: Under assumptions ( $i$ ) use a technique similar to the one in the proof of proposition 16.8; under assumptions (ii) follow the same strategy as in exercise 16.7.]

## 17 Globalization by Line-Search

There is no guarantee that the local algorithms in chapters 14 and 15 will converge when they are started at a point $x_{1}$ far from a solution $x_{*}$ to problem $\left(P_{E}\right)$ or $\left(P_{E I}\right)$. They can generate erratic sequences, which may by chance enter the neighborhood of a solution and then converge to it; but most often, the sequences will not converge. There exist several ways of damping this uncoordinated behavior and modifying the computation of the iterates so as to force their convergence. Two classes of techniques can be distinguished among them: line-search and trust-region. The former is presented in this chapter.

In methods with line-search, the iterates are generated by the recurrence

$$
x_{k+1}=x_{k}+\alpha_{k} d_{k},
$$

where $d_{k}$ is a direction in $\mathbb{R}^{n}$ and $\alpha_{k}>0$ is a stepsize, computed by a linesearch technique (see chapter 3), whose aim is to decrease a merit function. In this chapter, we consider algorithms in which $d_{k}$ solves or approximately solves the osculating quadratic program (14.8)/(15.4) of the Newton/SQP algorithm in chapters $14 / 15$ and the merit function is the function $\Theta_{\sigma}$ in chapter 16 . For convenience, we recall the definition of $\Theta_{\sigma}$ :

$$
\begin{equation*}
\Theta_{\sigma}(x)=f(x)+\sigma\left\|c(x)^{\#}\right\|_{P} \tag{17.1}
\end{equation*}
$$

where $\|\cdot\|_{P}$ denotes an arbitrary norm and the notation $(\cdot)^{\#}$ was introduced on page 194. Properties of function $\Theta_{\sigma}$ are studied in chapter 16 ; remember that this function is usually nondifferentiable.

Let us stress the originality of this approach, which uses the solution to the osculating quadratic program to minimize $\Theta_{\sigma}$. If $d_{k}$ were an arbitrary descent direction of the nondifferentiable merit function $\Theta_{\sigma}$, for example the steepest-descent direction, the resulting algorithm would not necessarily converge (see § 9.2.1). We shall show, however, that the difficulty coming from nonsmoothness does not occur if the search direction $d_{k}$ solves the osculating quadratic problem (15.4). As for the stepsize, the value $\alpha_{k}=1$ is preferred, in order to preserve the quadratic convergence of the local method. We shall see that the unit stepsize is actually accepted when $x_{k}$ is close to a strong solution to $\left(P_{E I}\right)$, provided some modifications of the search direction or the merit function are made. Therefore, the final algorithm can also be viewed as
a quadratically convergent method for minimizing the structured nonsmooth function $\Theta_{\sigma}$, a speed of convergence that cannot be obtained with general purpose nondifferentiable algorithms like those presented in part II of this book.

The concept of exactness plays an important part in the success of the approach we have just outlined. Without this property, it might indeed have been necessary to adapt $\sigma$ continually to make the solution $d_{k}$ to the quadratic problem a descent direction of the merit function $\Theta_{\sigma}$. This is illustrated for an equality constraint problem in figure 17.1 (a single constraint and two


Fig. 17.1. Importance of exactness: $\sigma$ too small (l), giving descent (m), giving exactness (r)
variables). The figure provides three pictures showing the level curves of $\Theta_{\sigma}$ for three increasing values of $\sigma\left(\bar{x}_{\sigma}\right.$ is the minimizer of $\left.\Theta_{\sigma}\right)$. They also show the constraint manifold (the bold curve at the bottom) and the Newton direction at $x_{k}$ (the arrow). We assume that the current iterate $x_{k}$ is close to $x_{*}$ (hence the figure gives greatly enlarged views) and that the multiplier $\lambda_{k}$ is also close to $\lambda_{*}$, so that the Newton direction $d_{k}$ points towards $x_{*}$ (this is a consequence of the quadratic convergence result in chapter 14). We can see that $d_{k}$ is an ascent direction of $\Theta_{\sigma}$ if $\sigma$ is not large enough (left-hand picture). In this case, there is no hope in finding a positive stepsize $\alpha_{k}$ along $d_{k}$ that provides a decrease in $\Theta_{\sigma}$. In the middle picture, $\sigma$ is large enough to make $d_{k}$ a descent direction of $\Theta_{\sigma}$, although not large enough to make $\Theta_{\sigma}$ exact at $x_{*}$. In the right-hand picture, the penalty parameter $\sigma$ is large enough to have $\bar{x}_{\sigma}=x_{*}$ (exactness of $\Theta_{\sigma}$ ) and this gives $d_{k}$ a greater chance of being a descent direction of $\Theta_{\sigma}$. As we shall see, other conditions must also be satisfied. Observe finally that the nondifferentiability of $\Theta_{\sigma}$ manifests itself in the pictures by the lack of smoothness of its level curves when they cross the constraint manifold.

To get descent property of $d_{k}$, it will be necessary to increase $\sigma$ at some iterations, but the exactness property of $\Theta_{\sigma}$ for a finite value of $\sigma$ will allow the algorithm to do this finitely often. This is a very desirable property, which makes the proof of convergence possible. As soon as $\sigma$ is fixed, $\Theta_{\sigma}$
plays the role of an immutable reference, which is able to appreciate the progress towards the solution, whatever may happen to the iterates.

This chapter describes and analyzes two classes of algorithms. Line-search SQP algorithms (§17.1) are based on the SQP direction of chapter 15 and use line-search on $\Theta_{\sigma}$ to enforce its convergence. We derive conditions that ensure the descent property of the SQP direction on $\Theta_{\sigma}$ and study the global convergence of the algorithm. This analysis assumes the strict convexity of the osculating quadratic program defining the SQP direction (as well as its feasibility), which may require not using the Hessian of the Lagrangian, but a positive definite approximation thereof (chapter 18 explains how to generate quasi-Newton approximations). The truncated SQP algorithm of $\S 17.2$ is presented as a line-search method that can use the exact Hessian of the Lagrangian (although we restrict the analysis to equality constrained problems). In this case, it is the way to solve the quadratic program approximately (discarding tangent negative curvature information) that allows the algorithm to generate descent directions of the merit function $\Theta_{\sigma}$. The so-called Maratos effect (nonadmissibility of the unit stepsize asymptotically) is discussed in $\S 17.3$, and the most common remedies for this phenomenon are described.

### 17.1 Line-Search SQP Algorithms

The quadratic program (QP) considered in this section is slightly more general than (15.4): the Hessian of the Lagrangian $L\left(x_{k}, \lambda_{k}\right)$ is replaced by some $n \times n$ symmetric matrix $M_{k}$. This allows us to include the Newton and the quasi-Newton versions of SQP in the same framework. On the other hand, the descent property of the QP solution and convergence of the line-search SQP algorithm often require the positive definiteness of $M_{k}$. The osculating quadratic problem in $d$ becomes:

$$
\left\{\begin{array}{l}
\min _{d} \nabla f\left(x_{k}\right)^{\top} d+\frac{1}{2} d^{\top} M_{k} d  \tag{17.2}\\
c_{E}\left(x_{k}\right)+A_{E}\left(x_{k}\right) d=0 \\
c_{I}\left(x_{k}\right)+A_{I}\left(x_{k}\right) d \leq 0
\end{array}\right.
$$

A stationary point $d_{k}$ of this QP satisfies, for some multiplier $\lambda_{k}^{\mathrm{QP}} \in \mathbb{R}^{m}$, the optimality conditions:

$$
\left\{\begin{array}{l}
(a) \nabla f_{k}+M_{k} d_{k}+A_{k}^{\top} \lambda_{k}^{\mathrm{QP}}=0  \tag{17.3}\\
(b)\left(c_{k}+A_{k} d_{k}\right)^{\#}=0 \\
(c)\left(\lambda_{k}^{\mathrm{QP}}\right)_{I} \geq 0 \\
(d)\left(\lambda_{k}^{\mathrm{QP}}\right)_{I}^{\top}\left(c_{k}+A_{k} d_{k}\right)_{I}=0
\end{array}\right.
$$

For short, we have set $\nabla f_{k}=\nabla f\left(x_{k}\right), c_{k}=c\left(x_{k}\right)$, and $A_{k}=A\left(x_{k}\right)=c^{\prime}\left(x_{k}\right)$.
Let us now outline the line-search SQP algorithm that uses $\Theta_{\sigma}$ as a merit function. The description includes references to numerical techniques, whose
sense will be clarified further in the section. The analysis of this algorithm is the subject of this section.

## LINE-SEARCH SQP:

Choose an initial iterate $\left(x_{1}, \lambda_{1}\right) \in \mathbb{R}^{n} \times \mathbb{R}^{m}$.
Compute $f\left(x_{1}\right), c\left(x_{1}\right), \nabla f\left(x_{1}\right)$, and $A_{1}:=c^{\prime}\left(x_{1}\right)$.
Set $k=1$.

1. Stop if the KKT conditions (13.1) holds at $\left(x_{*}, \lambda_{*}\right) \equiv\left(x_{k}, \lambda_{k}\right)$ (optimality is reached).
2. Compute a symmetric matrix $M_{k}$, approximating the Hessian of the Lagrangian, and find a primal-dual stationary point $\left(d_{k}, \lambda_{k}^{\mathrm{QP}}\right)$ of the quadratic problem (17.2) (i.e., a solution to the optimality conditions (17.3)), which is assumed to be feasible.
3. Adapt $\sigma_{k}$ if necessary (the update rule must satisfy (17.9) to ensure convergence, but a rule similar to the one on page 295 is often used).
4. Choose $\alpha_{k}>0$ along $d_{k}$ so as to obtain a "sufficient" decrease in $\Theta_{\sigma_{k}}$ (for example, use the line-search technique given on page 296).
5. Set $x_{k+1}:=x_{k}+\alpha_{k} d_{k}$ and update $\lambda_{k} \rightarrow \lambda_{k+1}$.
6. Compute $\nabla f\left(x_{k+1}\right)$ and $A_{k+1}:=c^{\prime}\left(x_{k+1}\right)$.
7. Increase $k$ by 1 and go to 1 .

This algorithm does not specify how to update the dual variables $\lambda_{k}$. Some authors do a line-search on $\lambda$ with the help of a primal-dual merit function, which therefore involves $\lambda$-values. Others compute $\lambda_{k+1}$ from $x_{k+1}$ as in the primal algorithm of $\S 14.3$. Another possibility is also to take

$$
\begin{equation*}
\lambda_{k+1}:=\lambda_{k}+\alpha_{k}\left(\lambda_{k}^{\mathrm{QP}}-\lambda_{k}\right), \tag{17.4}
\end{equation*}
$$

where $\alpha_{k}$ is the stepsize used for the primal variables. It has already been said that the role of $\lambda_{k}$ is less important than that of $x_{k}$, because it intervenes in the algorithm only through the matrix $M_{k}$ (for example the Hessian of the Lagrangian) in (17.2). The few requirements on the way the new multiplier is determined reflects in some way this fact.

General assumptions for this section. We assume throughout this section that $f$ and $c$ are differentiable in an open set containing the segments $\left[x_{k}, x_{k+1}\right]$ that link the successive iterates. We also assume that the quadratic problem (17.2) is always feasible (i.e., its constraints are compatible).

In practice, the last assumption on the feasibility of (17.2) is far from always being satisfied at each iteration. Therefore, carefully written codes
use techniques and heuristics for dealing with infeasible quadratic programs. For more computational efficiency, it is also often better to have a different penalty factor associated with each constraint, as in exercise 16.8. For simplicity, we keep a merit function with a single penalty parameter $\sigma$, knowing that an extension is possible without difficulty.

## Decrease in $\Theta_{\sigma}$ Along $d_{k}$

The merit function $\Theta_{\sigma}$ decreases from $x_{k}$ along $d_{k}$ if $d_{k}$ is a descent direction of $\Theta_{\sigma}$ at $x_{k}$ (we saw in lemma 16.4 that $\Theta_{\sigma}$ has directional derivatives), meaning that

$$
\Theta_{\sigma}^{\prime}\left(x_{k} ; d_{k}\right)<0
$$

We focus on this issue in this subsection.
The next proposition identifies three conditions that make $d_{k}$ a descent direction of $\Theta_{\sigma}: \sigma$ is large enough, $M_{k}$ is positive definite, and $x_{k}$ is not a stationary point of $\left(P_{E I}\right)$. Such a result is useful for the quasi-Newton versions of SQP, where the positive definiteness of $M_{k}$ is preserved. To hold, the result needs the following assumption on the norm $\|\cdot\|_{P}$ used in $\Theta_{\sigma}$ :

$$
\begin{equation*}
v \mapsto\left\|v^{\#}\right\|_{P} \text { is convex. } \tag{17.5}
\end{equation*}
$$

This hypothesis is weaker than (16.15) (see exercise 17.1).
Proposition 17.1 (descent property). If $\left(d_{k}, \lambda_{k}^{\mathrm{QP}}\right)$ satisfies the optimality conditions (17.3) and if $\|\cdot\|_{P}$ satisfies (17.5), then

$$
\begin{equation*}
\Theta_{\sigma}^{\prime}\left(x_{k} ; d_{k}\right) \leq \nabla f_{k}^{\top} d_{k}-\sigma\left\|c_{k}^{\#}\right\|_{P}=-d_{k}^{\top} M_{k} d_{k}+\left(\lambda_{k}^{Q \mathrm{P}}\right)^{\top} c_{k}-\sigma\left\|c_{k}^{\#}\right\|_{P} \tag{17.6}
\end{equation*}
$$

If, in addition, $\sigma \geq\left\|\lambda_{k}^{\mathrm{QP}}\right\|_{D}$, we have

$$
\Theta_{\sigma}^{\prime}\left(x_{k} ; d_{k}\right) \leq-d_{k}^{\top} M_{k} d_{k}
$$

Hence $\Theta_{\sigma}^{\prime}\left(x_{k} ; d_{k}\right)<0$, if $\sigma \geq\left\|\lambda_{k}^{\mathrm{QP}}\right\|_{D}$, if $M_{k}$ is positive definite, and if $x_{k}$ is not a stationary point of problem $\left(P_{E I}\right)$.

Proof. Since a norm has directional derivatives and is Lipschitzian (like any convex function), the function $v \rightarrow\left\|v^{\#}\right\|_{P}$ has directional derivatives. From (17.5) and $(17.3)_{b}$, we have for $\left.t \in\right] 0,1[$ :

$$
\begin{aligned}
\left\|\left(c_{k}+t A_{k} d_{k}\right)^{\#}\right\|_{P} & =\left\|\left[(1-t) c_{k}+t\left(c_{k}+A_{k} d_{k}\right)\right]^{\#}\right\|_{P} \\
& \leq(1-t)\left\|c_{k}^{\#}\right\|_{P}+t\left\|\left(c_{k}+A_{k} d_{k}\right)^{\#}\right\|_{P} \\
& =(1-t)\left\|c_{k}^{\#}\right\|_{P} .
\end{aligned}
$$

Therefore

$$
\left(\|\cdot \#\|_{P}\right)^{\prime}\left(c_{k} ; A_{k} d_{k}\right)=\lim _{t \rightarrow 0+} \frac{1}{t}\left(\left\|\left(c_{k}+t A_{k} d_{k}\right)^{\#}\right\|_{P}-\left\|c_{k}^{\#}\right\|_{P}\right) \leq-\left\|c_{k}^{\#}\right\|_{P}
$$

Then, with $(17.3)_{a},(17.3)_{b}$ and $(17.3)_{d}$, we prove (17.6):

$$
\begin{aligned}
\Theta_{\sigma}^{\prime}\left(x_{k} ; d_{k}\right) & \leq \nabla f_{k}^{\top} d_{k}-\sigma\left\|c_{k}^{\#}\right\|_{P} \\
& =-d_{k}^{\top} M_{k} d_{k}-\left(\lambda_{k}^{\mathrm{QP}}\right)^{\top} A_{k} d_{k}-\sigma\left\|c_{k}^{\#}\right\|_{P} \\
& =-d_{k}^{\top} M_{k} d_{k}+\left(\lambda_{k}^{\mathrm{QP}}\right)^{\top} c_{k}-\sigma\left\|c_{k}^{\#}\right\|_{P}
\end{aligned}
$$

If $\sigma \geq\left\|\lambda_{k}^{Q \mathrm{P}}\right\|_{D}$, using $(17.3)_{c}$ and the generalized Cauchy-Schwarz inequality (16.14), we have

$$
\left(\lambda_{k}^{\mathrm{QP}}\right)^{\top} c_{k}-\sigma\left\|c_{k}^{\#}\right\|_{P} \leq\left(\lambda_{k}^{Q \mathrm{P}}\right)^{\top} c_{k}^{\#}-\sigma\left\|c_{k}^{\#}\right\|_{P} \leq\left(\left\|\lambda_{k}^{\mathrm{QP}}\right\|_{D}-\sigma\right)\left\|c_{k}^{\#}\right\|_{P} \leq 0
$$

Now, the second inequality of the proposition is obtained from (17.6). If $\Theta_{\sigma}^{\prime}\left(x_{k} ; d_{k}\right)=0$ and $M_{k}$ is positive definite, then $d_{k}=0$. From (17.3), it follows that $x_{k}$ is stationary, with $\lambda_{k}^{\mathrm{QP}}$ as its associated multiplier.

Note that equality holds in (17.6) if there are only equality constraints (see the proof of lemma 17.4 below), but this is not necessarily the case when $I \neq \emptyset$ (this is the subject of exercise 17.2). Therefore, algorithms requiring the computation of $\Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)$ often use the negative upper bound given by the right-hand side of (17.6):

$$
\begin{equation*}
\Delta_{k}:=\nabla f_{k}^{\top} d_{k}-\sigma_{k}\left\|c_{k}^{\#}\right\|_{P}=-d_{k}^{\top} M_{k} d_{k}+\left(\lambda_{k}^{\mathrm{QP}}\right)^{\top} c_{k}-\sigma_{k}\left\|c_{k}^{\#}\right\|_{P} \tag{17.7}
\end{equation*}
$$

We have indexed $\sigma$ by $k$, since its value will have to be modified at some iterations.

## Update of the Penalty Parameter $\sigma_{k}$

A consequence of proposition 17.1 is that when $x_{k}$ is nonstationary, when $M_{k}$ is positive definite, and when $\sigma_{k}$ satisfies

$$
\begin{equation*}
\sigma_{k}>\left\|\lambda_{k}^{Q \mathrm{P}}\right\|_{D} \tag{17.8}
\end{equation*}
$$

then $\Delta_{k}<0$ and the solution $d_{k}$ to the osculating quadratic problem is a descent direction of $\Theta_{\sigma_{k}}$ at $x_{k}$, meaning that $\Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)<0$. Inequality (17.8) reminds us of the exactness condition $\sigma>\left\|\lambda_{*}\right\|_{D}$ found for $\Theta_{\sigma}$ in chapter 16 and is therefore natural: by maintaining (17.8) at each iteration, the algorithm ensures the exactness of $\Theta_{\sigma}$ at convergence ( $\sigma_{k}=\sigma$ for large $k$ and $\lambda_{k}^{\mathrm{QP}} \rightarrow \lambda_{*}$ ).

To maintain (17.8) at each iteration, it is necessary to modify $\sigma_{k}$ sometimes (the evolution of $\lambda_{k}^{\mathrm{QP}}$ cannot be known when the algorithm is started). Global convergence will show that this inequality has to be imposed with some safeguard, given by the positive constant $\bar{\sigma}$ below. To keep some generality, we shall just specify the properties that an adequate adaptation rule for $\sigma_{k}$ must enjoy:

$$
\left\{\begin{array}{l}
(a) \sigma_{k} \geq\left\|\lambda_{k}^{\mathrm{QP}}\right\|_{D}+\bar{\sigma}, \quad \text { for all } k \geq 1, \\
\text { (b) there exists an index } k_{1} \text { such that: }  \tag{17.9}\\
\text { if } k \geq k_{1} \text { and } \sigma_{k-1} \geq\left\|\lambda_{k}^{\mathrm{QP}}\right\|_{D}+\bar{\sigma}, \text { then } \sigma_{k}=\sigma_{k-1}, \\
\text { (c) if }\left\{\sigma_{k}\right\} \text { is bounded, } \sigma_{k} \text { is modified finitely often. }
\end{array}\right.
$$

Property (a) means that a little more than (17.8) must hold at each iteration. With (b), we assume that, after finitely many steps, $\sigma_{k-1}$ is modified only when necessary, to obtain $(a)$. Finally, $(c)$ requires that each modification of $\sigma_{k}$ is significant, so as to stabilize the sequence $\left\{\sigma_{k}\right\}$ : asymptotically, the merit function should no longer depend on the iteration index.

It can be checked that the following rule, proposed by Mayne and Polak [250], satisfies these properties (the constant 1.5 is given to be specific; actually, any constant $>1$ is appropriate):

$$
\begin{aligned}
& \text { if } \quad \sigma_{k-1} \geq\left\|\lambda_{k}^{\mathrm{QP}}\right\|_{D}+\bar{\sigma} \\
& \text { then } \quad \sigma_{k}=\sigma_{k-1} \\
& \text { else } \quad \sigma_{k}=\max \left(1.5 \sigma_{k-1},\left\|\lambda_{k}^{\mathrm{QP}}\right\|_{D}+\bar{\sigma}\right) .
\end{aligned}
$$

Having a large parameter $\sigma_{k}$ is harmless for the theoretical convergence, but can be disastrous in practice; so it must sometimes be decreased. In this case, the properties in (17.9) may no longer be satisfied and convergence may no longer be guaranteed. Nevertheless, an update rule like the one below is often used (the constants 1.1 and 1.5 can be replaced by any constant $>1$ ):

$$
\begin{aligned}
& \text { UPDATE RULE FOR } \sigma_{k} \text { : } \\
& \text { if } \sigma_{k-1} \geq 1.1\left(\left\|\lambda_{k}^{Q \mathrm{P}}\right\|_{D}+\bar{\sigma}\right), \\
& \text { then } \quad \sigma_{k}=\left(\sigma_{k-1}+\left\|\lambda_{k}^{Q \mathrm{P}}\right\|_{D}+\bar{\sigma}\right) / 2 \\
& \text { else } \text { if } \quad \sigma_{k-1} \geq\left\|\lambda_{k}^{Q P}\right\|_{D}+\bar{\sigma}, \\
& \text { then } \sigma_{k}=\sigma_{k-1} . \\
& \quad \text { else } \quad \sigma_{k}=\max \left(1.5 \sigma_{k-1},\left\|\lambda_{k}^{Q P}\right\|_{D}+\bar{\sigma}\right)
\end{aligned}
$$

In this rule, when the previous penalty factor $\sigma_{k-1}$ exceeds 1.1 times the minimal threshold $\left\|\lambda_{k}^{\mathrm{QP}}\right\|_{D}+\bar{\sigma}$, the new factor $\sigma_{k}$ is set to the arithmetic mean of this threshold and of $\sigma_{k-1}$.

It is often better to use a different penalty factor for each constraint (in particular, when the constraints have very different orders of magnitude). This is done by taking as a penalty function $\Theta_{\sigma}(x)=f(x)+\left\|S c(x)^{\#}\right\|_{P}$, where $S=\operatorname{Diag}\left(\sigma_{1}, \ldots, \sigma_{n}\right)$. The case of the $\ell_{1}$ norm is considered in exercise 16.8.

## Line-Search

The determination of the stepsize $\alpha_{k}>0$ along $d_{k}$, forcing the decrease in $\Theta_{\sigma_{k}}$, must be done in a precise manner (see $\S 3$ for unconstrained problems). We shall enforce satisfaction of the following Armijo condition [12]: $\omega$ being a fixed constant in $] 0, \frac{1}{2}[$, one determines $\alpha>0$ such that

$$
\begin{equation*}
x_{k}+\alpha d_{k} \in \Omega \quad \text { and } \quad \Theta_{\sigma_{k}}\left(x_{k}+\alpha d_{k}\right) \leq \Theta_{\sigma_{k}}\left(x_{k}\right)+\omega \alpha \Delta_{k} \tag{17.10}
\end{equation*}
$$

The requirement $\omega<\frac{1}{2}$ comes from the necessity of having asymptotic admissibility of the unit stepsize (see §17.3); it is essential neither for consistency of (17.10) nor for global convergence ( $\omega \in] 0,1[$ would be sufficient). The value of $\Delta_{k}$ in (17.10) should ideally be $\Theta_{\sigma_{k}}^{\prime}\left(x_{k}, d_{k}\right)$, but since this directional derivative is not easy to compute, we take the negative upper bound given by (17.7).

Since $\Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right) \leq \Delta_{k}<0$ and $\omega<1$, one can easily verify that it is possible to find $\alpha_{k}>0$ satisfying (17.10). However, this Armijo condition does not eliminate unduly small $\alpha_{k}$ 's, which might impair convergence of the iterates to a stationary point. This explains the following line-search algorithm. A constant $\left.\beta \in] 0, \frac{1}{2}\right]$ is chosen.

## BACKTRACKING LINE-SEARCH:

Set $i=0, \alpha_{k, 0}=1$.

1. If (17.10) is satisfied with $\alpha=\alpha_{k, 0}$, set $\alpha_{k}=\alpha$ and exit.
2. Choose $\alpha_{k, i+1} \in\left[\beta \alpha_{k, i},(1-\beta) \alpha_{k, i}\right]$.

3 . Increase $i$ by 1 and go to 1 .

Taking for example $\beta=\frac{1}{2}$, the stepsize selected by this algorithm is the first element encountered in the list $\left\{1, \frac{1}{2}, \frac{1}{4}, \frac{1}{8}, \cdots\right\}$ satisfying (17.10). Taking the first of these stepsizes does prevent $\alpha$ from being too small. The determination of $\alpha_{k, i+1}$ in the interval $\left[\beta \alpha_{k, i},(1-\beta) \alpha_{k, i}\right]$ should be done using interpolation formulas.

## Global Convergence with Positive Definite Hessian Approximations

In this subsection, we analyze the global convergence of the line-search SQP algorithm given on page 292 , when $\sigma_{k}$ is adapted by a rule satisfying properties (17.9), the stepsize $\alpha_{k}$ is determined by the line-search algorithm on page 296 , and the matrices $M_{k}$ used in the osculating quadratic program (17.2) are maintained positive definite, in such a way that

$$
\begin{equation*}
\left\{M_{k}\right\} \text { and }\left\{M_{k}^{-1}\right\} \text { are bounded. } \tag{17.11}
\end{equation*}
$$

This is a strong assumption. For example, it is not known whether it is satisfied in the quasi-Newton versions of SQP. Besides, if $M_{k}=L\left(x_{k}, \lambda_{k}\right)$, positive definiteness is not guaranteed. We shall, however, accept this assumption, which allows a simple convergence proof.

Theorem 17.2 (global convergence of the line-search SQP algorithm). Suppose that $f$ and $c$ are of class $C^{1,1}$ in $\Omega$ and that $\left\|\cdot{ }^{\#}\right\|_{P}$ is
convex. Consider the line-search SQP algorithm on page 292, using symmetric positive definite matrices $M_{k}$ satisfying (17.11) and an update rule of $\sigma_{k}$ satisfying (17.9). Then, starting the algorithm at a point $x_{1} \in \Omega$, one of the following situations occurs:
(i) the sequence $\left\{\sigma_{k}\right\}$ is unbounded, in which case $\left\{\lambda_{k}^{\mathrm{QP}}\right\}$ is also unbounded;
(ii) there exists an index $k_{2}$ such that $\sigma_{k}=\sigma$ for $k \geq k_{2}$, and at least one of the following situations occurs:
(a) $\Theta_{\sigma}\left(x_{k}\right) \rightarrow-\infty$,
(b) $\operatorname{dist}\left(x_{k}, \Omega^{c}\right) \rightarrow 0$,
(c) $\nabla_{x} \ell\left(x_{k}, \lambda_{k}^{\mathrm{QP}}\right) \rightarrow 0, c_{k}^{\#} \rightarrow 0,\left(\lambda_{k}^{\mathrm{QP}}\right)_{I} \geq 0$ and $\left(\lambda_{k}^{\mathrm{QP}}\right)_{I}^{\top}\left(c_{k}\right)_{I} \rightarrow 0$.

Proof. If $\left\{\sigma_{k}\right\}$ is unbounded, we see from rule $(17.9)_{b}$ that $\left\{\lambda_{k}^{\mathrm{QP}}: \sigma_{k} \neq \sigma_{k-1}\right\}$ is unbounded. If $\left\{\sigma_{k}\right\}$ is bounded, rule $(17.9)_{c}$ shows that there exists an index $k_{2}$ such that $\sigma_{k}=\sigma$ for all $k \geq k_{2}$. It remains to show that one of the situations (ii-a), (ii-b), or (ii-c) occurs. For this, we suppose that (ii-a) and (ii-b) do not hold and show (ii-c).

Each iteration after $k_{2}$ forces the decrease in the same function $\Theta_{\sigma}$. Since $\Theta_{\sigma}\left(x_{k}\right) \geq C>-\infty$, Armijo's condition (17.10) shows that

$$
\alpha_{k} \Delta_{k} \rightarrow 0
$$

Then, if we show $\alpha_{k} \geq \underline{\alpha}>0$, the result (ii-c) will follow. Indeed, from $\Delta_{k} \rightarrow 0,(17.6)$ and $(17.9)_{a}$, we deduce

$$
d_{k}^{\top} M_{k} d_{k} \rightarrow 0 \quad \text { and } \quad c_{k}^{\#} \rightarrow 0
$$

Because $M_{k}$ is positive definite and has a bounded inverse, $d_{k} \rightarrow 0$. Then, from $(17.3)_{a}$ and the boundedness of $M_{k}$, we see that $\nabla_{x} \ell\left(x_{k}, \lambda_{k}^{\mathrm{QP}}\right) \rightarrow 0$. On the other hand, $(17.3)_{c}$ shows that $\left(\lambda_{k}^{\mathrm{QP}}\right)_{I} \geq 0$. Finally, $\Delta_{k}=\nabla f_{k}^{\top} d_{k}-$ $\sigma\left\|c_{k}^{\#}\right\|_{P} \rightarrow 0$ and $c_{k}^{\#} \rightarrow 0$ imply that $\nabla f_{k}^{\top} d_{k} \rightarrow 0$ and, using $(17.3)_{a}$, $\left(\lambda_{k}^{\mathrm{QP}}\right)^{\top} A_{k} d_{k} \rightarrow 0$. Hence, from $(17.3)_{d}$ and $(17.3)_{b}$,

$$
\begin{aligned}
\left(\lambda_{k}^{\mathrm{QP}}\right)_{I}^{\top}\left(c_{k}\right)_{I} & =-\left(\lambda_{k}^{\mathrm{QP}}\right)_{I}^{\top}\left(A_{k} d_{k}\right)_{I} \\
& =\left(\lambda_{k}^{\mathrm{QP}}\right)_{E}^{\top}\left(A_{k} d_{k}\right)_{E}+o(1) \\
& =-\left(\lambda_{k}^{\mathrm{QP}}\right)_{E}^{\top}\left(c_{k}\right)_{E}+o(1) \\
& =o(1),
\end{aligned}
$$

because $\left\{\lambda_{k}^{\mathrm{QP}}\right\}$ is bounded and $\left(c_{k}\right)_{E} \rightarrow 0$.
Therefore, it remains to prove that $\alpha_{k} \geq \underline{\alpha}>0$, for all $k$ and some constant $\underline{\alpha}$. We can consider the indices $k$ of $\mathcal{K}:=\left\{k \geq k_{2}: \alpha_{k}<1\right\}$. Then from the rule determining the stepsize, $\alpha_{k} \in\left[\beta \bar{\alpha}_{k},(1-\beta) \bar{\alpha}_{k}\right]$ for some $\left.\left.\bar{\alpha}_{k} \in\right] 0,1\right]$ satisfying

$$
\alpha_{k}+\bar{\alpha}_{k} d_{k} \notin \Omega \quad \text { or } \quad \Theta_{\sigma}\left(x_{k}+\bar{\alpha}_{k} d_{k}\right)>\Theta_{\sigma}\left(x_{k}\right)+\omega \bar{\alpha}_{k} \Delta_{k}
$$

For large $k$, the first condition is impossible because $d_{k} \rightarrow 0$ would then imply that $\operatorname{dist}\left(x_{k}, \Omega^{c}\right) \rightarrow 0$. Hence, for large $k \in \mathcal{K}$, we have

$$
\begin{equation*}
\Theta_{\sigma}\left(x_{k}+\bar{\alpha}_{k} d_{k}\right)>\Theta_{\sigma}\left(x_{k}\right)+\omega \bar{\alpha}_{k} \Delta_{k} \tag{17.12}
\end{equation*}
$$

Let us expand the left-hand side of (17.12). Using the smoothness of $f$ and $c$, $\bar{\alpha}_{k} \leq 1$, the convexity of $\left\|\cdot{ }^{\#}\right\|_{P}$ (hence its Lipschitz continuity), (17.3) ${ }_{b}$, and finally (17.6)-(17.7), we have successively

$$
\begin{aligned}
f\left(x_{k}+\bar{\alpha}_{k} d_{k}\right) & =f_{k}+\bar{\alpha}_{k} \nabla f_{k}^{\top} d_{k}+O\left(\bar{\alpha}_{k}^{2}\left\|d_{k}\right\|^{2}\right) \\
c\left(x_{k}+\bar{\alpha}_{k} d_{k}\right) & =c_{k}+\bar{\alpha}_{k} A_{k} d_{k}+O\left(\bar{\alpha}_{k}^{2}\left\|d_{k}\right\|^{2}\right) \\
& =\left(1-\bar{\alpha}_{k}\right) c_{k}+\bar{\alpha}_{k}\left(c_{k}+A_{k} d_{k}\right)+O\left(\bar{\alpha}_{k}^{2}\left\|d_{k}\right\|^{2}\right) \\
\left\|c\left(x_{k}+\bar{\alpha}_{k} d_{k}\right)^{\#}\right\|_{P} & \leq\left(1-\bar{\alpha}_{k}\right)\left\|c_{k}^{\#}\right\|_{P}+\bar{\alpha}_{k}\left\|\left(c_{k}+A_{k} d_{k}\right)^{\#}\right\|_{P}+O\left(\bar{\alpha}_{k}^{2}\left\|d_{k}\right\|^{2}\right) \\
& =\left(1-\bar{\alpha}_{k}\right)\left\|c_{k}^{\#}\right\|_{P}+O\left(\bar{\alpha}_{k}^{2}\left\|d_{k}\right\|^{2}\right) \\
\Theta_{\sigma}\left(x_{k}+\bar{\alpha}_{k} d_{k}\right) & \leq \Theta_{\sigma}\left(x_{k}\right)+\bar{\alpha}_{k} \Delta_{k}+C_{1} \bar{\alpha}_{k}^{2}\left\|d_{k}\right\|^{2} .
\end{aligned}
$$

Then (17.12) yields

$$
-(1-\omega) \bar{\alpha}_{k} \Delta_{k} \leq C_{1} \bar{\alpha}_{k}^{2}\left\|d_{k}\right\|^{2}
$$

But $\Delta_{k}=-d_{k}^{\top} M_{k} d_{k}+\left(\lambda_{k}^{\mathrm{QP}}\right)^{\top} c_{k}-\sigma\left\|c_{k}^{\#}\right\|_{P} \leq-d_{k}^{\top} M_{k} d_{k} \leq-C_{2}\left\|d_{k}\right\|^{2}$ (boundedness of $\left\{M_{k}^{-1}\right\}$ ), so that we deduce from the above inequality:

$$
\bar{\alpha}_{k} \geq\left(C_{2} / C_{1}\right)(1-\omega)>0
$$

because $\omega<1$. The positive lower bound on $\alpha_{k}$ can therefore be taken as $\underline{\alpha}:=\beta\left(C_{2} / C_{1}\right)(1-\omega)$. This concludes the proof.

Among the situations described in theorem 17.2, only situation (ii-c) is satisfactory. In this case, every cluster point of $\left\{\left(x_{k}, \lambda_{k}^{Q P}\right)\right\}$ satisfies the optimality conditions (KKT). Unfortunately, any of the other situations may occur. For example, $(i)$ may occur in the example in figure 16.1 when $\left\{x_{k}\right\}$ converges to $x_{*}^{\prime}$, a point where $\lambda_{*}$ is not defined. Situation (ii-a) will occur if, outside of the feasible set, $f$ decreases more rapidly than $\left\|c(\cdot)^{\#}\right\|_{P}$ increases, and if $x_{1}$ is taken far enough from the feasible set; the example

$$
\min \left\{-x^{2}: x=0\right\}
$$

with $\|\cdot\|_{P}=|\cdot|$, is such. Finally, situation (ii-b) occurs if $\Omega$ contains no stationary point.

### 17.2 Truncated SQP

In this section, we consider another globalization technique of the Newton algorithm to solve the problem with only equality constraints:

$$
\left(P_{E}\right) \quad\left\{\begin{array}{l}
\min _{x} f(x) \\
c(x)=0
\end{array}\right.
$$

The local algorithm was introduced in § 14.1 and we refer the reader to $\S 14.4$ (in the subsection entitled "The reduced system approach") for the notation. In contrast to the approach used in the previous section, we do not replace here the Hessian of the Lagrangian by a positive definite approximation. This was useful to ensure the well-posedness of the osculating quadratic program and the decrease in $\Theta_{\sigma}$ along the computed direction. Instead, we describe an algorithm that directly exploits the curvature of the problem (i.e., the second derivatives of $f$ and $c$ ) gathered in the Hessian of the Lagrangian, even in the presence of nonconvexity.

Here also, the computed direction will be a descent direction of the merit function $\Theta_{\sigma}$, which allows global convergence. Therefore, it must differ from Newton's direction, but the modification only needs to be done at points where the reduced Hessian of the Lagrangian is not positive definite. This form of weak nonconvexity can therefore be detected by the algorithm, which is a nice feature. The idea is similar to the truncated Newton algorithm in unconstrained optimization (see §6.4): the truncated conjugate gradient (CG) algorithm is used to solve, sometimes approximately, the reduced linear system (see (14.32))

$$
\begin{equation*}
H_{k} u_{k}=v_{k} \tag{17.13}
\end{equation*}
$$

where

$$
\begin{equation*}
H_{k}:=Z_{k}^{-\top} L_{k} Z_{k}^{-} \quad \text { and } \quad v_{k}:=-g_{k}+Z_{k}^{-\top} L_{k} A_{k}^{-} c_{k} . \tag{17.14}
\end{equation*}
$$

Note that the reduced Hessian of the Lagrangian $H_{k}$ is symmetric but may be indefinite. By the truncated CG, the algorithm aims at collecting only the "positive definite part" of $H_{k}$. This is obtained by stopping the CG iterations certainly before a conjugate direction $w$ is a negative curvature direction for $H_{k}$ (more precisely, before $w^{\top} H_{k} w$ becomes less than an appropriate positive threshold). Let us denote by $\tilde{u}_{k}$ the approximate solution to (17.13) computed by the truncated CG algorithm. We shall show that the search direction

$$
\begin{equation*}
d_{k}=-A_{k}^{-} c_{k}+Z_{k}^{-} \tilde{u}_{k} \tag{17.15}
\end{equation*}
$$

is then a descent direction of $\Theta_{\sigma}$ provided $\sigma$ is larger than an easily computable threshold. Another interesting property of this approach is that, since $H_{k}$ is positive definite around a strong solution to $\left(P_{E}\right)$, the CG iterations can be pursued up to completion close to such a solution, so that local quadratic convergence is not prevented.

Let us look at this in more detail.

## Truncated CG Iterations

The truncated conjugate gradient (TCG) algorithm to solve (17.13) is presented below. For clarity, we drop the index $k$ of the Newton algorithm and
denote by $i$ the CG iteration index (in superscript). For $i=0, \ldots, j$, Algorithm TCG generates iterates $u^{i}$, approximating the solution to (17.13), residuals $r^{i}:=H u^{i}-v$, and conjugate directions $w^{i}$. The algorithm can be stopped at any iteration (global convergence of the truncated SQP method will not be affected by this), but it must certainly be interrupted at $u^{j}$ if the next conjugate direction $w^{j}$ is a quasi-negative curvature direction for $H$. This means that the following inequality does not hold with $i=j$ :

$$
\begin{equation*}
\left(w^{i}\right)^{\top} H w^{i} \geq \nu\left\|w^{i}\right\|_{2}^{2} \tag{17.16}
\end{equation*}
$$

The threshold $\nu>0$ is assumed to be independent of the index $k$, although an actual implementation would use a more sophisticated rule for setting this parameter, allowing small values when approaching a solution. Hence, Algorithm TCG simply discards quasi-negative directions. It is in this way that nonconvexity is dealt with.

## Algorithm TCG For (17.13):

1. Choose $\nu>0$. Set $u^{0}=0$ and $r^{0}=-v$, where $v$ is defined by (17.14).
2. For $i=0,1, \ldots$ do the following:
2.1. If desired or if $r^{i}=0$, stop to iterate and go to step 3 with $j=i$.
2.2. Compute a new conjugate direction:

$$
w^{i}=\left\{\begin{array}{lr}
-r^{i} & \text { if } i=0 \\
-r^{i}+\frac{\left\|r^{i}\right\|^{2}}{\left\|r^{i-1}\right\|^{2}} w^{i-1} & \text { if } i \geq 1
\end{array}\right.
$$

2.3. Compute $p^{i}=H w^{i}$.
2.4. If (17.16) does not hold, go to step 3 with $j=i$.
2.5. Compute the new iterate $u^{i+1}=u^{i}+t^{i} w^{i}$ and the new residual $r^{i+1}=r^{i}+t^{i} p^{i}$, with the stepsize

$$
t^{i}=\frac{\left\|r^{i}\right\|^{2}}{\left(w^{i}\right)^{\top} p^{i}}
$$

3. Take as the approximate solution to (17.13):

$$
\tilde{u}= \begin{cases}v & \text { if } j=0 \\ u^{j} & \text { if } j \geq 1\end{cases}
$$

Observe that, since the first iterate of Algorithm TCG is $u^{0}=0$, the first CG direction is $w^{0}=-r^{0}=v$, the right-hand side of (17.13). This is important for the analysis that follows. Another key point is that the directions $w^{i}$ are conjugate: $w^{i_{1}} H w^{i_{2}}=0$ for $i_{1} \neq i_{2}$. Note finally that Algorithm TCG chooses to output the approximate solution $u^{j}$ currently obtained when $j \geq 1$ (it is different from zero), but $\tilde{u}=w^{0}=v$ when $j=0\left(u^{0}=0\right.$ in this case $)$.

Lemma 17.3. The vector $\tilde{u}$ computed by Algorithm TCG has the form

$$
\begin{equation*}
\tilde{u}=J v, \tag{17.17}
\end{equation*}
$$

where $J$ is the identity matrix when $j=0$ and

$$
\begin{equation*}
J=\sum_{i=0}^{j-1} \frac{w^{i}\left(w^{i}\right)^{\top}}{\left(w^{i}\right)^{\top} H w^{i}} \tag{17.18}
\end{equation*}
$$

when $j \geq 1$. Furthermore $\|J\|_{2} \leq \max \left(1, \frac{j}{\nu}\right)$.
Proof. If $i=0, u=v$ and the result follows. Otherwise Algorithm TCG generates conjugate directions $w^{0}, \ldots, w^{j-1}$. By orthogonality of $r^{i}$ and $w^{i-1}$, by the fact that the algorithm starts with $u^{0}=0$, and by conjugacy of the directions $w^{i}$, one has for $1 \leq i \leq j$ :

$$
\begin{aligned}
\left\|r^{i}\right\|^{2} & =-\left(w^{i}\right)^{\top} r^{i} \\
& =-\left(w^{i}\right)^{\top}\left(H u^{i}-v\right) \\
& =-\left(w^{i}\right)^{\top} H\left(\sum_{l=0}^{i-1} t^{l} w^{l}\right)+\left(w^{i}\right)^{\top} v \\
& =\left(w^{i}\right)^{\top} v
\end{aligned}
$$

Also, $\left\|r^{0}\right\|^{2}=\left(w^{0}\right)^{\top} v$. Therefore

$$
\tilde{u}=\sum_{i=0}^{j-1} t^{i} w^{i}=\sum_{i=0}^{j-1} \frac{\left(w^{i}\right)^{\top} v}{\left(w^{i}\right)^{\top} H w^{i}} w^{i}=\left(\sum_{i=0}^{j-1} \frac{w^{i}\left(w^{i}\right)^{\top}}{\left(w^{i}\right)^{\top} H w^{i}}\right) v
$$

This proves (17.18).
The upper bound on $\|J\|_{2}$ comes from the fact that $\left\|v v^{\top}\right\|_{2}=\|v\|_{2}^{2}$ and (17.16).

Note that, when $j \geq 1$, the matrix $J$ is positive semi-definite with rank $j$. In view of (17.13) and (17.17), this matrix appears as a kind of "pseudoinverse of the positive definite part" of $H$.

## Descent Property

In the next lemma, we give conditions ensuring that the direction $d_{k}$ given by (17.15) is a descent direction of $\Theta_{\sigma_{k}}$. For this, it is convenient to give another expression of $d_{k}$ by introducing the following right inverse of $A_{k}$ :

$$
\begin{equation*}
\widetilde{A}_{k}^{-}:=\left(I-Z_{k}^{-} J_{k} Z_{k}^{-\top} L_{k}\right) A_{k}^{-} \tag{17.19}
\end{equation*}
$$

This is the right inverse $\widehat{A}_{k}^{-}$in (14.33), in which $H_{k}^{-1}$ has been substituted by its approximation $J_{k}$. Then

$$
\begin{equation*}
d_{k}=\tilde{r}_{k}+\tilde{t}_{k} \tag{17.20}
\end{equation*}
$$

where

$$
\tilde{r}_{k}=-\tilde{A}_{k}^{-} c_{k} \quad \text { and } \quad \tilde{t}_{k}=-Z_{k}^{-} J_{k} g_{k}
$$

We also use the multiplier associated with $\widetilde{A}_{k}^{-}$:

$$
\begin{equation*}
\tilde{\lambda}_{k}=-\widetilde{A}_{k}^{-\top} \nabla f_{k} . \tag{17.21}
\end{equation*}
$$

How to compute this multiplier efficiently is dealt with in the next subsection.
Lemma 17.4 (descent property). Suppose that $f$ and $c$ are differentiable at $x_{k}$. Let $d_{k}$ be given by (17.15), where $\tilde{u}_{k}$ is the approximate solution to (17.13) computed by Algorithm TCG. Then $\Theta_{\sigma_{k}}$ has a directional derivative in the direction $d_{k}$, whose value is given by

$$
\begin{equation*}
\Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)=-g_{k}^{\top} J g_{k}+\tilde{\lambda}_{k}^{\top} c_{k}-\sigma_{k}\left\|c_{k}\right\|_{P} \tag{17.22}
\end{equation*}
$$

It is negative if $x_{k}$ is nonstationary and $\sigma_{k}>\left\|\tilde{\lambda}_{k}\right\|_{D}$.
Proof. Since a norm is Lipschitz continuous and has directional derivatives, $\|\cdot\|_{P} \circ c$ has directional derivatives (see lemma 13.1). Using the fact that $d_{k}$ satisfies the linearized constraints (i.e., $A_{k} d_{k}=-c_{k}$ ), one has $\left(\|\cdot\|_{P} \circ c\right)^{\prime}\left(x_{k} ; d_{k}\right)=\left(\|\cdot\|_{P}\right)^{\prime}\left(c_{k} ;-c_{k}\right)=-\left\|c_{k}\right\|_{P}$. Therefore

$$
\Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)=\nabla f_{k}^{\top} d_{k}-\sigma_{k}\left\|c_{k}\right\|_{P}
$$

Using (17.20) and (17.21), we get (17.22).
Suppose now that $\sigma_{k}>\left\|\tilde{\lambda}_{k}\right\|_{D}$. Since $\tilde{\lambda}_{k}^{\top} c_{k} \leq\left\|\tilde{\lambda}_{k}\right\|_{D}\left\|c_{k}\right\|_{P}$, we obtain

$$
\Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right) \leq-g_{k}^{\top} J_{k} g_{k}+\left(\left\|\tilde{\lambda}_{k}\right\|_{D}-\sigma_{k}\right)\left\|c_{k}\right\|_{P} \leq 0
$$

If $\Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)=0$, it follows that $c_{k}=0$ and $g_{k}^{\top} J_{k} g_{k}=0$. If the number of CG iterations $j_{k}=0$, then $J_{k}=I$, hence $g_{k}=0$ and $x_{k}$ is stationary. It remains to show that $j_{k}$ cannot be $\geq 1$ when $\Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)=0$. If $j_{k} \geq 1$, one would have $v_{k} \neq 0$ (see step 2.1 of Algorithm TCG) and therefore $g_{k} \neq 0$ (since $c_{k}=0$ ). But with the structure of $J_{k}$ and the fact that $w_{k}^{0}=v_{k}=-g_{k}$ when $c_{k}=0$, one would have $g_{k}^{\top} J_{k} g_{k} \geq\left(g_{k}^{\top} w_{k}^{0}\right)^{2} /\left(\left(w_{k}^{0}\right)^{\top} H_{k} w_{k}^{0}\right)=\left\|g_{k}\right\|^{4} /\left(g_{k}^{\top} H g_{k}\right)>0$, which would contradict the fact that $g_{k}^{\top} J_{k} g_{k}=0$.

## Computation of $\tilde{\lambda}_{\boldsymbol{k}}$

Let us drop the index $k$. From (17.21) and (17.19), the definition of $\tilde{\lambda}$ involves the matrix $J$ :

$$
\tilde{\lambda}=-A^{-\top}\left(\nabla f-L Z^{-} J g\right)
$$

We do not want to store this matrix, however. In fact, to compute $\tilde{\lambda}$, one has to evaluate $\bar{u}=J g$, which is the approximate solution to

$$
\begin{equation*}
H \bar{u}=g, \tag{17.23}
\end{equation*}
$$

obtained by using the same conjugate directions $w^{i}$ and the same products $p^{i}=H w^{i}, i=0, \ldots, j-1$, as those used to compute the approximate solution $\tilde{u}$ to (17.13) by Algorithm TCG. The computation of $\tilde{u}$ and $\bar{u}$ can be made in parallel, hence avoiding the need to store the conjugate directions $w^{i}$ (or $J$ ) or the need to compute twice the Hessian-vector products $p^{i}=H w^{i}$. This is what Algorithm TCG2 below does. Its outputs are $\tilde{u}$ and $\bar{u}$.

Algorithm TCG2 FOR (17.13) AND (17.23):

1. Choose $\nu>0$. Set $u^{0}=0, r^{0}=-v, \bar{u}^{0}=0$, and $\bar{r}^{0}=-g$, where $v$ is defined by (17.14).
2. For $i=0,1, \ldots$ do the following:
2.1. If desired or if $r^{i}=0$, stop to iterate and go to step 3 with $j=i$.
2.2. Compute a new conjugate direction:

$$
w^{i}=\left\{\begin{array}{lr}
-r^{i} & \text { if } i=0 \\
-r^{i}+\frac{\left\|r^{i}\right\|^{2}}{\left\|r^{i-1}\right\|^{2}} w^{i-1} & \text { if } i \geq 1
\end{array}\right.
$$

2.3. Compute $p^{i}=H w^{i}$.
2.4. If (17.16) does not hold, go to step 3 with $j=i$.
2.5. Compute the new iterates $u^{i+1}=u^{i}+t^{i} w^{i}$ and $\bar{u}^{i+1}=\bar{u}^{i}+$ $\bar{t}^{i} w^{i}$ and the new residuals $r^{i+1}=r^{i}+t^{i} p^{i}$ and $\bar{r}^{i+1}=\bar{r}^{i}+\bar{t}^{i} p^{i}$, with the stepsizes

$$
t^{i}=\frac{\left\|r^{i}\right\|^{2}}{\left(w^{i}\right)^{\top} p^{i}} \quad \text { and } \quad \bar{t}^{i}=-\frac{\left(\bar{r}^{i}\right)^{\top} w^{i}}{\left(w^{i}\right)^{\top} p^{i}}
$$

3. Take as the approximate solution to (17.13) and (17.23):

$$
\tilde{u}=\left\{\begin{array}{l}
v \text { if } j=0 \\
u^{j} \text { if } j \geq 1
\end{array} \quad \text { and } \quad \bar{u}=\left\{\begin{array}{l}
g \text { if } j=0 \\
\bar{u}^{j} \text { if } j \geq 1 .
\end{array}\right.\right.
$$

It may occur that the linear system (17.23) is solved before (17.13). In this case, the stepsizes $\bar{t}^{i}$ vanish and $\bar{u}^{i}$ is no longer modified. It is easy to verify that $\tilde{\lambda}$ is obtained from $\bar{u}$ by:

$$
\begin{equation*}
\tilde{\lambda}=-A^{-\top}\left(\nabla f-L Z^{-} \bar{u}\right) . \tag{17.24}
\end{equation*}
$$

Indeed, since $\bar{u}^{0}=0$, one has for $1 \leq i \leq j$ :

$$
\left(w^{i}\right)^{\top} \bar{r}^{i}=\left(w^{i}\right)^{\top}\left(H \bar{u}^{i}-g\right)=\left(w^{i}\right)^{\top} H\left(\sum_{l=0}^{i-1} \bar{t}^{l} w^{l}\right)-\left(w^{i}\right)^{\top} g=-\left(w^{i}\right)^{\top} g
$$

Hence

$$
\bar{u}=\sum_{i=0}^{j-1} \bar{t}^{i} w^{i}=\sum_{i=0}^{j-1} \frac{\left(w^{i}\right)^{\top} g}{\left(w^{i}\right)^{\top} H w^{i}} w^{i}=J g
$$

## The Truncated SQP Algorithm and its Global Convergence

The truncated SQP algorithm to solve problem $\left(P_{E}\right)$ generates a sequence $\left\{x_{k}\right\}_{k \geq 1}$ by the recurrence

$$
x_{k+1}=x_{k}+\alpha_{k} d_{k}
$$

where the direction $d_{k} \in \mathbb{R}^{n}$ is determined by (17.15), with $\tilde{u}_{k}$ computed by Algorithm TCG2, and the stepsize $\alpha_{k}>0$ is determined by a line-search on the merit function $\Theta_{\sigma_{k}}$.

According to lemma $17.4, d_{k}$ is a descent direction of $\Theta_{\sigma_{k}}$ provided $x_{k}$ is nonstationary and $\sigma_{k}>\left\|\tilde{\lambda}_{k}\right\|_{D}$. This requires a modification of $\sigma_{k}$ at some iterations and we assume that a rule respecting conditions similar to (17.9) is adopted: for some fixed constant $\bar{\sigma}>0$, the following holds

$$
\left\{\begin{array}{l}
\text { (a) } \sigma_{k} \geq\left\|\tilde{\lambda}_{k}\right\|_{D}+\bar{\sigma}, \text { for all } k \geq 1,  \tag{17.25}\\
\text { (b) there exists an index } k_{1} \text { such that: } \\
\text { if } k \geq k_{1} \text { and } \sigma_{k-1} \geq\left\|\tilde{\lambda}_{k}\right\|_{D}+\bar{\sigma}, \text { then } \sigma_{k}=\sigma_{k-1}, \\
\text { (c) if }\left\{\sigma_{k}\right\} \text { is bounded, } \sigma_{k} \text { is modified finitely often. }
\end{array}\right.
$$

Since at a nonstationary iterate $x_{k}, d_{k}$ is a descent direction of $\Theta_{\sigma_{k}}$, one can determine a stepsize $\alpha_{k}>0$ such that the following Armijo inequality holds

$$
\begin{equation*}
\Theta_{\sigma_{k}}\left(x_{k}+\alpha_{k} d_{k}\right) \leq \Theta_{\sigma_{k}}\left(x_{k}\right)+\omega \alpha_{k} \Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right) \tag{17.26}
\end{equation*}
$$

where $\omega$ is a constant chosen in $] 0, \frac{1}{2}[$. As in the line-search SQP algorithm on page 292, the stepsize is determined in step 4 below by backtracking.

We can now summarize the overall TSQP algorithm to solve the equality constrained problem $\left(P_{E}\right)$.

## Algorithm TSQP:

Choose an initial iterate $\left(x_{1}, \lambda_{1}\right) \in \mathbb{R}^{n} \times \mathbb{R}^{m}$.
Compute $f\left(x_{1}\right), c\left(x_{1}\right), \nabla f\left(x_{1}\right)$, and $A\left(x_{1}\right)$.
Set the constants $\nu>0$ (quasi-negative curvature threshold), $\omega \in$ ]0, $\frac{1}{2}$ [ (slope modifier in the Armijo condition), $\bar{\sigma}>0$ (penalty parameter threshold), and $\left.\beta \in] 0, \frac{1}{2}\right]$ (backtracking safeguard parameter).
Set $k=1$.

1. Stopping test: Stop if $c_{k}=0$ and $g_{k}=0$.
2. Step computation:

- Compute the restoration step $r_{k}=-A_{k}^{-} c_{k}$.
- Compute the reduced gradient $g_{k}=Z_{k}^{-\top} \nabla f_{k}$ and the right-hand side of (17.13) $v_{k}=-g_{k}-Z_{k}^{-\top} L_{k} r_{k}$.
- Run Algorithm TCG2 to compute $\tilde{u}_{k}$ and $\bar{u}_{k}$.
- Compute the full step $d_{k}=r_{k}+Z_{k}^{-} \tilde{u}_{k}$ and the multiplier $\tilde{\lambda}_{k}$ by (17.24).

3. Penalty parameter setting: Update $\sigma_{k}$ such that (17.25) holds.
4. Backtracking line-search:

- Set $\alpha=1$.
- While $\alpha$ does not satisfy Armijo's inequality (17.26), pick a new stepsize $\alpha$ in $[\beta \alpha,(1-\beta) \alpha]$.
- Set $\alpha_{k}=\alpha$.

5. New iterates: Set $x_{k+1}=x_{k}+\alpha_{k} d_{k}$ and $\lambda_{k+1}=\lambda_{k+1}^{\mathrm{LS}}$.

6 . Increase $k$ by 1 and go to 1 .

Before proving the global convergence of this algorithm, let us make some observations. In a practical algorithm, the stopping test in step 1 would be replaced by a condition checking that $c_{k}$ and $g_{k}$ are sufficiently small. In practice, in step 4 , the new stepsize chosen in the interval $[\beta \alpha,(1-\beta) \alpha]$ during the line-search should be obtained by interpolation. In step 5 , we have set the new multiplier $\lambda_{k+1}$ to the least-squares multiplier

$$
\lambda_{k}^{\mathrm{LS}}:=-A_{k}^{-\top} \nabla f_{k} .
$$

This makes Algorithm TSQP close to the primal version of Newton's algorithm analyzed in theorem 14.5. Another possibility would have been to choose $\lambda_{k+1}=\tilde{\lambda}_{k}$. Observe however that, even if the CG iterations of Algorithm TCG2 solve (17.13) and (17.23) exactly, $\tilde{\lambda}_{k} \neq \lambda_{k}^{Q \mathrm{PP}}$ (in this case $\tilde{\lambda}_{k}=\widehat{\lambda}_{k}$ given by (14.36), compare with (14.35)), so that with that choice of $\lambda_{k+1}$, Algorithm TSQP does not reduce to Newton's algorithm in a neighborhood of a strong solution.

Theorem 17.5 (global convergence of the line-search truncated SQP algorithm). Suppose that the functions $f$ and $c$ are twice continuously differentiable with Lipschitz continuous first derivatives. Suppose also that the sequences $\left\{\nabla f_{k}\right\},\left\{L_{k}\right\},\left\{A_{k}^{-}\right\}$, and $\left\{Z_{k}^{-}\right\}$generated by Algorithm TSQP are bounded. Then the sequence of penalty parameters $\left\{\sigma_{k}\right\}$ is stationary for sufficiently large $k: \sigma_{k}=\sigma$. If furthermore $\left\{\Theta_{\sigma}\left(x_{k}\right)\right\}$ is bounded below, the sequences $\left\{c_{k}\right\}$ and $\left\{g_{k}\right\}$ converge to 0 .

Proof. We denote by $C_{1}, C_{2}, \ldots$ positive constants, independent of $k$. We can assume that $\left\|c_{k}\right\|+\left\|g_{k}\right\|>0$ for all $k \geq 1$, because otherwise the conclusion is clear.

Note first, that the assumptions imply the boundedness of $\left\{\tilde{\lambda}_{k}\right\}$ (use (17.24), the boundedness of $\left\{A_{k}^{-}\right\},\left\{\nabla f_{k}\right\},\left\{L_{k}\right\},\left\{Z_{k}^{-}\right\}$, and that of $\left\{J_{k}\right\}$ given by lemma 17.3). Then by $(17.25)_{b},\left\{\sigma_{k}\right\}$ is also bounded, hence stationary for large enough $k$ (use (17.25) $)_{c}$ ). From Armijo's inequality (17.26), $\Theta_{\sigma}\left(x_{k}\right)$ is decreasing. It is also bounded below (by assumption), hence it converges. This implies that $\alpha_{k} \Theta_{\sigma}^{\prime}\left(x_{k} ; d_{k}\right)$ tends to 0 , or equivalently (use lemma 17.4 and $(17.25)_{a}$ )

$$
\begin{equation*}
\alpha_{k} g_{k}^{\top} J_{k} g_{k} \rightarrow 0 \quad \text { and } \quad \alpha_{k} c_{k} \rightarrow 0 \tag{17.27}
\end{equation*}
$$

Let us now show that $\left\{\alpha_{k}\right\}$ is bounded away from 0 . From the linesearch (step 4), when $\alpha_{k}<1$, there is a stepsize $\left.\left.\bar{\alpha}_{k} \in\right] 0,1\right]$ such that $\alpha_{k} \in\left[\beta \bar{\alpha}_{k},(1-\beta) \bar{\alpha}_{k}\right]$ and

$$
\Theta_{\sigma}\left(x_{k}+\bar{\alpha}_{k} d_{k}\right)>\Theta_{\sigma}\left(x_{k}\right)+\omega \bar{\alpha}_{k} \Theta_{\sigma}^{\prime}\left(x_{k} ; d_{k}\right) .
$$

Using the smoothness of $f$ and $c$ and the fact that $d_{k}$ satisfies the linearized constraints, one has successively

$$
\begin{aligned}
f\left(x_{k}+\bar{\alpha}_{k} d_{k}\right) & =f\left(x_{k}\right)+\bar{\alpha}_{k} f^{\prime}\left(x_{k}\right) \cdot d_{k}+O\left(\bar{\alpha}_{k}^{2}\left\|d_{k}\right\|^{2}\right), \\
c\left(x_{k}+\bar{\alpha}_{k} d_{k}\right) & =\left(1-\bar{\alpha}_{k}\right) c\left(x_{k}\right)+O\left(\bar{\alpha}_{k}^{2}\left\|d_{k}\right\|^{2}\right), \\
\Theta_{\sigma}\left(x_{k}+\bar{\alpha}_{k} d_{k}\right) & \leq \Theta_{\sigma}\left(x_{k}\right)+\bar{\alpha}_{k} \Theta_{\sigma}^{\prime}\left(x_{k} ; d_{k}\right)+C_{1} \bar{\alpha}_{k}^{2}\left\|d_{k}\right\|^{2} .
\end{aligned}
$$

Therefore $(\omega-1) \Theta_{\sigma}^{\prime}\left(x_{k} ; d_{k}\right)<C_{1} \bar{\alpha}_{k}\left\|d_{k}\right\|^{2}$ or

$$
\begin{equation*}
g_{k}^{\top} J_{k} g_{k}+\left\|c_{k}\right\|_{P}<C_{2} \bar{\alpha}_{k}\left\|d_{k}\right\|^{2} \tag{17.28}
\end{equation*}
$$

where $C_{2}=C_{1} /((1-\omega) \min (1, \bar{\sigma}))$. With the boundedness of $\left\{A_{k}^{-}\right\},\left\{Z_{k}^{-}\right\}$, $\left\{L_{k}\right\}$, and $\left\{J_{k}\right\}$, we have $d_{k}=O\left(\left\|J_{k}^{1 / 2} v_{k}\right\|+\left\|c_{k}\right\|_{P}\right)$ and, due to the form of $v_{k}, d_{k}=O\left(\left\|J_{k}^{1 / 2} g_{k}\right\|+\left\|c_{k}\right\|_{P}\right)$. Then, inequality (17.28) becomes

$$
g_{k}^{\top} J_{k} g_{k}+\left\|c_{k}\right\|_{P}<C_{3} \bar{\alpha}_{k}\left(g_{k}^{\top} J_{k} g_{k}+\left\|c_{k}\right\|_{P}^{2}\right)
$$

From (17.27), $\alpha_{k} c_{k} \rightarrow 0$ and therefore for large $k$

$$
g_{k}^{\top} J_{k} g_{k}<C_{3} \bar{\alpha}_{k} g_{k}^{\top} J_{k} g_{k}
$$

This inequality shows that $g_{k}^{\top} J_{k} g_{k} \neq 0$ when $\alpha_{k}<1$ and $k$ is large enough and that $\left\{\bar{\alpha}_{k}\right\}$ is bounded away from zero. Since $\alpha_{k} \geq \beta \bar{\alpha}_{k},\left\{\alpha_{k}\right\}$ is also bounded away from zero.

From (17.27)

$$
\begin{equation*}
g_{k}^{\top} J_{k} g_{k} \rightarrow 0 \quad \text { and } \quad c_{k} \rightarrow 0 \tag{17.29}
\end{equation*}
$$

It remains to show that $g_{k} \rightarrow 0$. Assume the opposite: there is a constant $\gamma>0$ and subsequence $\mathcal{K}$ such that $\left\|g_{k}\right\| \geq \gamma$ for $k \in \mathcal{K}$. Using the first term of the expression (17.18) of $J_{k}$ when $j_{k} \geq 1, w_{k}^{0}=v_{k}$, and the boundedness of $\left\{H_{k}\right\}$, one can write

$$
g_{k}^{\top} J_{k} g_{k} \geq \min \left(\left\|g_{k}\right\|_{2}^{2}, \frac{\left(g_{k}^{\top} v_{k}\right)^{2}}{v_{k}^{\top} H_{k} v_{k}}\right) \geq \min \left(\gamma^{2}, C_{4} \frac{\left(g_{k}^{\top} v_{k}\right)^{2}}{\left\|v_{k}\right\|^{2}}\right)
$$

The numerator can be bounded below as follows:

$$
\begin{aligned}
\left(g_{k}^{\top} v_{k}\right)^{2} & =\left[-\left\|g_{k}\right\|^{2}+O\left(\left\|g_{k}\right\|\left\|c_{k}\right\|\right)\right]^{2} \\
& =\left\|g_{k}\right\|^{4}+O\left(\left\|g_{k}\right\|^{3}\left\|c_{k}\right\|\right)+O\left(\left\|g_{k}\right\|^{2}\left\|c_{k}\right\|^{2}\right) \\
& \geq \frac{1}{2}\left\|g_{k}\right\|^{4}-C_{5}\left\|g_{k}\right\|^{2}\left\|c_{k}\right\|^{2} \\
& \geq\left\|g_{k}\right\|^{2}\left(\frac{1}{2} \gamma^{2}-C_{5}\left\|c_{k}\right\|^{2}\right),
\end{aligned}
$$

which is positive for large $k$ in $\mathcal{K}$. For the denominator, we use the upper bound:

$$
\left\|v_{k}\right\|^{2} \leq 2\left\|g_{k}\right\|^{2}+C_{6}\left\|c_{k}\right\|^{2} \leq\left\|g_{k}\right\|^{2}\left(2+C_{6}\left\|c_{k}\right\|^{2} / \gamma^{2}\right)
$$

Therefore for large $k$ in $\mathcal{K}$ :

$$
g_{k}^{\top} J_{k} g_{k} \geq \min \left(\gamma^{2}, \frac{\frac{1}{2} \gamma^{2}-C_{5}\left\|c_{k}\right\|^{2}}{2+C_{6}\left\|c_{k}\right\|^{2} / \gamma^{2}}\right)
$$

This is in contradiction with (17.29).

### 17.3 From Global to Local

In this section, we analyze conditions under which the line-search algorithms of the present chapter can transform themselves into the "local" algorithms of chapter 14. In view of the quadratic convergence of the local methods, this "mutation" is highly desirable. Because the direction generated by the local algorithm is used as a descent direction of some merit function, this transformation will occur if the line-search accepts the unit stepsize during the last iterations. This property is referred to as the asymptotic admissibility of the unit stepsize. We shall see that it is not guaranteed without certain modifications of the algorithms, which are therefore crucial for their efficiency.

For simplicity, we assume in this section that the problem has only equality constraints:

$$
\left(P_{E}\right) \quad\left\{\begin{array}{l}
\min _{x} f(x) \\
c(x)=0
\end{array}\right.
$$

Since our study is asymptotic, assuming convergence of the sequence $\left\{\left(x_{k}\right.\right.$, $\left.\left.\lambda_{k}\right)\right\}$ to a primal-dual solution $\left(x_{*}, \lambda_{*}\right)$, this simplification amounts to assuming that the active constraints are identified after finitely many iterations, in which case problem $\left(P_{E I}\right)$ reduces locally to a problem with only equality constraints (theorem 15.2 tells us something about this).

## The Maratos Effect

The merit function $\Theta_{\sigma}$ introduced in $\S 16.4$ and defined by

$$
\Theta_{\sigma}(x)=f(x)+\sigma\|c(x)\|_{P}
$$

does not necessarily accept unit stepsizes asymptotically. This is known as the Maratos effect. We mean by this that when $d_{k}$ solves the quadratic problem

$$
\left\{\begin{array}{l}
\min _{d} \nabla f\left(x_{k}\right)^{\top} d+\frac{1}{2} d^{\top} M_{k} d  \tag{17.30}\\
c\left(x_{k}\right)+A\left(x_{k}\right) d=0
\end{array}\right.
$$

we may have

$$
\begin{equation*}
\Theta_{\sigma}\left(x_{k}+d_{k}\right)>\Theta_{\sigma}\left(x_{k}\right) \tag{17.31}
\end{equation*}
$$

however close to $\left(x_{*}, L_{*}\right)$ the current pair $\left(x_{k}, M_{k}\right)$ may be.
The following counter-example demonstrates this fact. There, the considered iterate $x_{k}$ is on the constraint manifold: $c\left(x_{k}\right)=0$. We have seen in proposition 17.1 that, if $\sigma_{k} \geq\left\|\lambda_{k}^{\mathrm{QP}}\right\|_{D}$ and $M_{k}$ is positive definite, $\Theta_{\sigma_{k}}$ decreases along the Newton direction $d_{k}$, which means that, for small stepsizes, the decrease in $f$ along $d_{k}$ compensates the increase in $\|c\|_{P}$. In the counter-example, this compensation not longer holds for stepsizes close to 1 .

Counter-example 17.6. Consider the problem on $\mathbb{R}^{2}$

$$
\left\{\begin{array}{l}
\min _{x}-x_{1}+\tau\left(x_{1}^{2}+x_{2}^{2}-1\right) \\
x_{1}^{2}+x_{2}^{2}-1=0
\end{array}\right.
$$

where $\tau \in \mathbb{R}$. Its unique solution is $x_{*}=(1,0)$ and the associated multiplier is $\lambda_{*}=\frac{1}{2}-\tau$. The Hessian of the Lagrangian at the solution is $L_{*}=I$.

Suppose now that the step $d$ at $x$ is given by the osculating quadratic problem, defined at a feasible point $x$ with the matrix $M=L_{*}=I$ :

$$
\left\{\begin{array}{l}
\min _{d}-d_{1}+\frac{1}{2}\|d\|_{2}^{2} \\
x^{\top} d=0
\end{array}\right.
$$

Its solution for $x=(\cos \theta, \sin \theta)$ lying on the constraint is

$$
d=\binom{\sin ^{2} \theta}{-\sin \theta \cos \theta}
$$

and $c(x+\alpha d)=\alpha^{2} \sin ^{2} \theta$. Hence, if $\|\cdot\|_{P}=|\cdot|$,

$$
\begin{aligned}
\Theta_{\sigma}(x) & =-\cos \theta \\
\Theta_{\sigma}(x+\alpha d) & =-\cos \theta-\alpha \sin ^{2} \theta+(\tau+\sigma) \alpha^{2} \sin ^{2} \theta
\end{aligned}
$$

Then $\Theta_{\sigma}(x+d)>\Theta_{\sigma}(x)$ whenever $\tau+\sigma>1$ (and $\theta \neq 0$ ). Because $\sigma \geq$ $\left|\lambda_{*}\right| \equiv\left|\frac{1}{2}-\tau\right|$ is needed to have an exact penalty, $\Theta_{\sigma}$ increases for a unit stepsize if $\tau>\frac{3}{4}$.

Figure 17.2 shows the level curves of $\Theta_{\sigma}$ around the solution for $\tau=1$ and $\sigma=0.6$, as well as the Newton step $d$ from an $x$ on the constraint manifold (the bold curve), rather close to the solution $(1,0)$. One clearly observes that $\Theta_{\sigma}(x+d)>\Theta_{\sigma}(x)$.


Fig. 17.2. Example with a Maratos effect

This phenomenon somehow reveals a discrepancy between $\Theta_{\sigma}$ and the osculating quadratic problem used to compute $d_{k}$. Since this model is good (it yields local quadratic convergence), the blame must be put on the merit function, or on the way in which it is used. In the rest of this section, we analyze different remedies for the Maratos effect and prove that they are effective close to a solution. The Maratos effect can also occur far from a solution and it is then more difficult to deal with. The first remedy consists in modifying the step $d_{k}$ by adding to it a small displacement, called a second order correction, that does not prevent quadratic convergence. Another possibility is to modify the merit function, which is considered next.

## Modification of the Step: Second Order Correction

Example 17.6 has shown that there are situations in which, even close to the solution, the increase in $\|c(\cdot)\|_{P}$ from $x_{k}$ to $x_{k}+d_{k}$ is not compensated by a decrease in $f$, resulting finally in an increase in $\Theta_{\sigma}$. The remedy for the Maratos effect presented in this subsection consists in adding to $d_{k}$ a small correcting step $e_{k} \in \mathbb{R}^{n}$, whose aim is to decrease $\|c(\cdot)\|_{P}$. This additional step is defined by

$$
\begin{equation*}
e_{k}=-A_{k}^{-} c\left(x_{k}+d_{k}\right) \tag{17.32}
\end{equation*}
$$

where $A_{k}^{-}$is some right inverse of the Jacobian matrix $A_{k}=c^{\prime}\left(x_{k}\right)$, which is assumed to be surjective. Hence, $e_{k}$ is a constraint-restoration step at $x_{k}+d_{k}$. Figure 17.3 shows the second order correction for counter-example 17.6: the small step $e$ from $x+d$ to $x+d+e$.


Fig. 17.3. Second order correction

One speaks of second-order correction because $c\left(x_{k}+d_{k}\right)=O\left(\left\|d_{k}\right\|^{2}\right)$ and therefore $e_{k}=O\left(\left\|d_{k}\right\|^{2}\right)$ is of order 2 in $d_{k}$. This modification of $d_{k}$ preserves a possible quadratic convergence since, assuming $x_{k}+d_{k}-x_{*}=O\left(\left\|x_{k}-x_{*}\right\|^{2}\right)$, we have

$$
x_{k}+d_{k}+e_{k}-x_{*}=\left(x_{k}+d_{k}-x_{*}\right)+e_{k}=O\left(\left\|x_{k}-x_{*}\right\|^{2}\right),
$$

owing to the preceding estimate of $e_{k}$ and to the fact that $d_{k} \sim\left(x_{k}-x_{*}\right)$ (lemma 13.5).

Because $e_{k}$ is computed by evaluating $c$ at a point different from $x_{k}$, it cannot be guaranteed that $d_{k}+e_{k}$ is a descent direction of $\Theta_{\sigma_{k}}$ at $x_{k}$. Therefore, a line-search along this direction may be impossible. The least expensive approach is then to determine a stepsize $\alpha_{k}>0$ along the arc

$$
\alpha \mapsto p_{k}(\alpha)=x_{k}+\alpha d_{k}+\alpha^{2} e_{k} .
$$

It has the descent direction $d_{k}$ as a tangent at $\alpha=0$ and visits $x_{k}+d_{k}+e_{k}$ for $\alpha=1$. The stepsize $\alpha_{k}$ can be computed in the same way as along $d_{k}$, forcing at each iteration the inequality

$$
\begin{equation*}
\Theta_{\sigma_{k}}\left(x_{k}+\alpha_{k} d_{k}+\alpha_{k}^{2} e_{k}\right) \leq \Theta_{\sigma_{k}}\left(x_{k}\right)+\omega \alpha_{k} \Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right) \tag{17.33}
\end{equation*}
$$

for some $\left.\left.\alpha_{k} \in\right] 0,1\right]$. It is easy to verify that this inequality can always be satisfied, provided $d_{k}$ is a descent direction of $\Theta_{\sigma_{k}}$ at $x_{k}$.

In the next proposition, we give conditions under which the unit stepsize $\alpha_{k}=1$ is accepted in (17.33) when $x_{k}$ is near a strong solution to $\left(P_{E}\right)$. Part of these conditions is related to the matrix $M_{k}$, which must satisfy (17.34). This condition is of the form $t_{k} \geq o\left(\left\|d_{k}\right\|^{2}\right)$, for some real numbers $t_{k}$, which means that there must exist a sequence of real numbers $\left\{s_{k}\right\}$, such that $t_{k} \geq s_{k}$ and $s_{k}=o\left(\left\|d_{k}\right\|^{2}\right)$ when $k \rightarrow \infty$. Observe that this condition is satisfied when $M_{k}$ is "large enough". This is not surprising, since then the tangent step is small (see remark 2 on page 235) and the total step $d_{k}$ is close to the restoration step, along which the unit stepsize is known to be accepted by the norm of the constraints (see exercise 17.4). Observe also that condition (17.34) is satisfied when $M_{k}$ is the Hessian of the Lagrangian (with convergent multipliers), which corresponds to Newton's method.

Proposition 17.7 (admissibility of the unit step-size with a second order correction). Suppose that $f$ and $c$ are of class $C^{2}$ in a neighborhood of a solution $x_{*}$ to ( $P_{E}$ ) satisfying the second-order sufficient conditions of optimality and at which $A_{*}=c^{\prime}\left(x_{*}\right)$ is surjective. Let $\left\{x_{k}\right\}$ be a sequence converging to $x_{*}$, let $d_{k}$ satisfy the first-order optimality conditions of the osculating quadratic problem (17.30), and let $e_{k}$ be defined by (17.32). Suppose also that

- $\left\{A_{k}^{-}\right\}$is bounded and $d_{k} \rightarrow 0$,
- the matrix $M_{k}$ used in the osculating quadratic problem (17.30) overestimates the Hessian of the augmented Lagrangian $L_{*}^{r}:=L_{*}+r A_{*}^{\top} A_{*}$, in the sense that

$$
\begin{equation*}
d_{k}^{\top}\left(M_{k}-L_{*}^{r}\right) d_{k} \geq o\left(\left\|d_{k}\right\|^{2}\right) \tag{17.34}
\end{equation*}
$$

where $r \geq 0$ is such that $L_{*}^{r}$ is positive definite (such an $r$ always exists under the assumptions already stated, see exercise 16.1),

- the penalty parameter $\sigma_{k}$ used in $\Theta_{\sigma_{k}}$ satisfies

$$
\begin{equation*}
\left\|\lambda_{k}^{Q P}\right\|_{D} \leq \sigma_{k} \leq \hat{\sigma} \tag{17.35}
\end{equation*}
$$

where $\lambda_{k}^{\mathrm{QP}}$ is a multiplier associated with the constraints of (17.30) and $\hat{\sigma}$ is a constant.
Then, for $\omega<\frac{1}{2}$ and large enough $k$, there holds

$$
\Theta_{\sigma_{k}}\left(x_{k}+d_{k}+e_{k}\right) \leq \Theta_{\sigma_{k}}\left(x_{k}\right)+\omega \Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)
$$

Proof. Despite the nondifferentiability of $\Theta_{\sigma_{k}}$, one can obtain an expansion of $\Theta_{\sigma_{k}}\left(x_{k}+d_{k}+e_{k}\right)$ with a precision of order $o\left(\left\|d_{k}\right\|^{2}\right)$. This one follows from an expansion of $f\left(x_{k}+d_{k}+e_{k}\right)$ and $c\left(x_{k}+d_{k}+e_{k}\right)$ about $x_{k}$. Using the smoothness assumptions on $f$ and $c$, the constraint in (17.30), the definition of $e_{k}$ in (17.32), the boundedness of $\left\{A_{k}^{-}\right\}$, and the optimality of $\left(x_{*}, \lambda_{*}\right)$, we have successively

$$
\begin{aligned}
c\left(x_{k}+d_{k}\right) & =c_{k}+A_{k} d_{k}+\frac{1}{2} c^{\prime \prime}\left(x_{*}\right) \cdot d_{k}^{2}+o\left(\left\|d_{k}\right\|^{2}\right) \\
& =\frac{1}{2} c^{\prime \prime}\left(x_{*}\right) \cdot d_{k}^{2}+o\left(\left\|d_{k}\right\|^{2}\right) \\
e_{k} & =O\left(\left\|c\left(x_{k}+d_{k}\right)\right\|\right) \\
& =O\left(\left\|d_{k}\right\|^{2}\right), \\
c\left(x_{k}+d_{k}+e_{k}\right) & =c\left(x_{k}+d_{k}\right)+A_{k} e_{k}+o\left(\left\|e_{k}\right\|\right) \\
& =o\left(\left\|d_{k}\right\|^{2}\right), \\
-A_{k}^{-\top} \nabla f_{k} & =\lambda_{*}-A_{k}^{-\top}\left(\nabla f_{k}+A_{k}^{\top} \lambda_{*}\right) \\
& =\lambda_{*}+o(1), \\
\nabla f_{k}^{\top} e_{k} & =-\left(A_{k}^{-\top} \nabla f_{k}\right)^{\top} c\left(x_{k}+d_{k}\right) \\
& =\lambda_{*}^{\top} c\left(x_{k}+d_{k}\right)+o\left(\left\|d_{k}\right\|^{2}\right) \\
& =\frac{1}{2} \lambda_{*}^{\top}\left(c^{\prime \prime}\left(x_{*}\right) \cdot d_{k}^{2}\right)+o\left(\left\|d_{k}\right\|^{2}\right), \\
f\left(x_{k}+d_{k}+e_{k}\right) & =f_{k}+\nabla f_{k}^{\top}\left(d_{k}+e_{k}\right)+\frac{1}{2} d_{k}^{\top} \nabla^{2} f\left(x_{*}\right) d_{k}+o\left(\left\|d_{k}\right\|^{2}\right) \\
& =f_{k}+\nabla f_{k}^{\top} d_{k}+\frac{1}{2} d_{k}^{\top} L_{*} d_{k}+o\left(\left\|d_{k}\right\|^{2}\right) .
\end{aligned}
$$

With these estimates, the boundedness of $\left\{\sigma_{k}\right\}$, and the fact that, when there are only equality constraints, the directional derivative of $\Theta_{\sigma_{k}}$ in the direction $d_{k}$ can be written $\Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)=\nabla f_{k}^{\top} d_{k}-\sigma_{k}\left\|c_{k}\right\|_{P}$ (see the proof of lemma 17.4), one gets

$$
\begin{align*}
& \Theta_{\sigma_{k}}\left(x_{k}+d_{k}+e_{k}\right)-\Theta_{\sigma_{k}}\left(x_{k}\right)-\omega \Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right) \\
& =\nabla f_{k}^{\top} d_{k}+\frac{1}{2} d_{k}^{\top} L_{*} d_{k}-\sigma_{k}\left\|c_{k}\right\|_{P}-\omega \Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)+o\left(\left\|d_{k}\right\|^{2}\right) \\
& =(1-\omega) \Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)+\frac{1}{2} d_{k}^{\top} L_{*} d_{k}+o\left(\left\|d_{k}\right\|^{2}\right) \tag{17.36}
\end{align*}
$$

We have to show that the right-hand side of (17.36) is nonpositive asymptotically.

Using the optimality conditions of (17.30), the Cauchy-Schwarz inequality (16.14), and the bounds in (17.35), the directional derivative $\Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)=$ $\nabla f_{k}^{\top} d_{k}-\sigma_{k}\left\|c_{k}\right\|_{P}$ can also be written

$$
\begin{equation*}
\Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)=-d_{k}^{\top} M_{k} d_{k}+\left(\lambda_{k}^{\mathrm{QP}}\right)^{\top} c_{k}-\sigma_{k}\left\|c_{k}\right\|_{P} \leq-d_{k}^{\top} M_{k} d_{k} \tag{17.37}
\end{equation*}
$$

Since $d_{k}^{\top} L_{*} d_{k} \leq d_{k}^{\top} L_{*}^{r} d_{k}$ for a nonnegative $r,(17.36)$ becomes with (17.37) and (17.34):

$$
\begin{aligned}
& \Theta_{\sigma_{k}}\left(x_{k}+d_{k}+e_{k}\right)-\Theta_{\sigma_{k}}\left(x_{k}\right)-\omega \Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right) \\
& \leq\left(\frac{1}{2}-\omega\right)\left(-d_{k}^{\top} M_{k} d_{k}\right)-\frac{1}{2} d_{k}^{\top}\left(M_{k}-L_{*}^{r}\right) d_{k}+o\left(\left\|d_{k}\right\|^{2}\right) \\
& \leq\left(\frac{1}{2}-\omega\right)\left(-d_{k}^{\top} M_{k} d_{k}\right)+o\left(\left\|d_{k}\right\|^{2}\right)
\end{aligned}
$$

For large $k$, the right-hand side is nonpositive since, by (17.34) and the positive definiteness of $L_{*}^{r}, d_{k}^{\top} M_{k} d_{k} \geq d_{k}^{\top} L_{*}^{r} d_{k}+o\left(\left\|d_{k}\right\|^{2}\right) \geq C\left\|d_{k}\right\|^{2}$, for some positive constant $C$ and large $k$.

The result of proposition 17.7 has many variants. It is usually easy to prove them by adapting the arguments used in the proof above (basically by cleverly combining Taylor expansions of an appropriate order). For example, one can avoid using the Hessian of the augmented Lagrangian by replacing condition (17.34) by

$$
d_{k}^{\top} P_{*}^{\top}\left(M_{k}-L_{*}\right) P_{*} d_{k} \geq o\left(\left\|d_{k}\right\|^{2}\right)+o\left(\left\|c_{k}\right\|\right)
$$

where $P_{*}$ denotes a projection operator on $N\left(A_{*}\right)$. The proof of this claim has been left as an exercise.

Computing the correction step $e_{k}$ can be time-consuming for some applications, since this requires a new evaluation of the constraints at $x_{k}+d_{k}$. When $x_{k}$ is far from a solution, this step can also be very large, perturbing uselessly the SQP step $d_{k}$. Therefore meticulous implementations of the line-search SQP algorithm usually have a test for deciding whether $e_{k}$ must be computed and the arc-search detailed above must be substituted for the less expensive line-search. Counter-example 17.6 has shown that the Maratos effect occurs when $x_{k}$ is on the constraint manifold. On the other hand, truncation of the unit stepsize is unlikely to occur in the neighborhood of a solution when the transversal part of the step prevails. To see this, observe that when $c$ has its values in $\mathbb{R}^{n}$, the unit stepsize is accepted along Newton's direction to solve $c(x)=0$ when one uses $x \mapsto\|c(x)\|_{P}$ as a merit function (see exercise 17.4). These observations suggest that there may be a danger of small stepsize only when the restoration step is small with respect to the tangent step. The next proposition confirms this viewpoint. It shows that the unit stepsize is accepted asymptotically for the iterations satisfying the inequality

$$
\begin{equation*}
\left\|r_{k}\right\| \geq C_{\mathrm{ME}}\left\|t_{k}\right\| \tag{17.38}
\end{equation*}
$$

where $C_{\mathrm{ME}}$ is a positive constant and $\|\cdot\|$ is an arbitrary norm. To write this inequality, we have decomposed the full step $d_{k}$ into $d_{k}=r_{k}+t_{k}$, where the restoration step is written $r_{k}=-A_{k}^{-} c_{k}$, for some right inverse $A_{k}^{-}$of $A_{k}$, and the tangent step $t_{k} \in R\left(Z_{k}^{-}\right)$satisfies $\nabla f_{k}^{\top} t_{k} \leq 0$.

Proposition 17.8 (admissibility of the unit step-size at restoration prevailing iterations). Suppose that $f$ and $c$ are of class $C^{1}$ in a neighborhood of a stationary point $x_{*}$ of $\left(P_{E}\right)$. Let $\left\{x_{k}\right\}$ be a sequence converging to $x_{*}$ and $d_{k}=r_{k}+t_{k}$, where $r_{k}=-A_{k}^{-} c\left(x_{k}\right)$ and $t_{k} \in R\left(Z_{k}^{-}\right)$ satisfies $\nabla f\left(x_{k}\right)^{\top} t_{k} \leq 0$. Suppose that $\left\{A_{k}^{-}\right\}$and $\left\{\sigma_{k}\right\}$ are bounded, that $\sigma_{k} \geq\left\|A_{k}^{-\top} \nabla f\left(x_{k}\right)\right\|_{D}+\bar{\sigma}$ for some constant $\bar{\sigma}>0$, and that $\omega<1$. Then, for large indices $k$ for which (17.38) holds with a positive constant $C_{M E}$, one has

$$
\Theta_{\sigma_{k}}\left(x_{k}+d_{k}\right) \leq \Theta_{\sigma_{k}}\left(x_{k}\right)+\omega \Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)
$$

Proof. Here, as we shall see, first-order expansions are sufficient. Using the fact that $d_{k}=O\left(\left\|r_{k}\right\|\right)$ for the considered indices, one has

$$
\begin{aligned}
f\left(x_{k}+d_{k}\right) & =f_{k}+\nabla f_{k}^{\top} d_{k}+o\left(\left\|r_{k}\right\|\right) \\
c\left(x_{k}+d_{k}\right) & =c_{k}+A_{k} d_{k}+o\left(\left\|r_{k}\right\|\right) \\
& =o\left(\left\|r_{k}\right\|\right) .
\end{aligned}
$$

Therefore, using $\Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)=\nabla f_{k}^{\top} d_{k}-\sigma_{k}\left\|c_{k}\right\|_{P}$ (see the proof of lemma 17.4), $\nabla f_{k}^{\top} t_{k} \leq 0, \omega<1, \nabla f_{k}^{\top} r_{k} \leq\left\|A_{k}^{-\top} \nabla f_{k}\right\|_{D}\left\|c_{k}\right\|_{P}$, and $r_{k}=O\left(\left\|c_{k}\right\|_{P}\right)$ :

$$
\begin{aligned}
& \Theta_{\sigma_{k}}\left(x_{k}+d_{k}\right)-\Theta_{\sigma_{k}}\left(x_{k}\right)-\omega \Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right) \\
& =(1-\omega) \nabla f_{k}^{\top} d_{k}-(1-\omega) \sigma_{k}\left\|c_{k}\right\|_{P}+o\left(\left\|r_{k}\right\|\right) \\
& \leq(1-\omega)\left(\left\|A_{k}^{-\top} \nabla f_{k}\right\|_{D}-\sigma_{k}\right)\left\|c_{k}\right\|_{P}+o\left(\left\|r_{k}\right\|\right) \\
& \leq-(1-\omega) \bar{\sigma}\left\|c_{k}\right\|_{P}+o\left(\left\|c_{k}\right\|_{P}\right)
\end{aligned}
$$

which is negative for large $k$.

A consequence of this result is that, optimization codes implementing the second order correction often decide to compute $e_{k}$ and to do an arcsearch, only at iterations where (17.38) does not hold. The constant $C_{\mathrm{ME}}$ is determined by heuristics.

## Modification of the Merit Function: <br> Nondifferentiable Augmented Lagrangian

Another way of getting the asymptotic admissibility of the unit stepsize is to change the merit function. Remember that $d_{k}$ is obtained by minimizing a quadratic model of the Lagrangian subject to linearized constraints. Hence, taking

$$
\ell_{\mu, \sigma}(x)=f(x)+\mu^{\top} c(x)+\sigma\|c(x)\|_{P}
$$

as a merit function should be convenient, insofar as $\mu$ is close enough to $\lambda_{*}$ and $\sigma$ is small enough. The validity of this intuition is confirmed by proposition 17.9 below.

Beforehand, observe that the problem

$$
\left\{\begin{array}{l}
\min _{x} f(x)+\mu^{\top} c(x) \\
c(x)=0, \quad x \in \Omega
\end{array}\right.
$$

is clearly equivalent to $\left(P_{E}\right)$. Now, let $x_{*}$ be a solution to $\left(P_{E}\right)$, with associated multiplier $\lambda_{*}$. Then $x_{*}$ is still a solution to the problem above, with associated multiplier $\lambda_{*}-\mu$. Therefore, the results of $\S 16.4$ imply that $\ell_{\mu, \sigma}$ is exact if

$$
\sigma>\left\|\lambda_{*}-\mu\right\|_{D}
$$

On the other hand, one easily computes

$$
\ell_{\mu, \sigma}^{\prime}\left(x_{k} ; d_{k}\right)=-d_{k}^{\top} M_{k} d_{k}+\left(\lambda_{k}^{\mathrm{QP}}-\mu\right)^{\top} c_{k}-\sigma\left\|c_{k}\right\|_{P},
$$

which is therefore negative if $M_{k}$ is positive definite and

$$
\sigma \geq\left\|\lambda_{k}^{Q P}-\mu\right\|_{D}
$$

Figure 17.4 shows the level curves of $\ell_{\mu, \sigma}$ for counter-example 17.6 , with


Fig. 17.4. Nondifferentiable augmented Lagrangian
$\tau=1, \mu=-0.55$, and $\sigma=0.1$.
Proposition 17.9 (admissibility of the unit step-size with a nondifferentiable augmented Lagrangian). Suppose that $f$ and $c$ are of class $C^{2}$ in a neighborhood of a solution $x_{*}$ to $\left(P_{E}\right)$, satisfying the second-order sufficient conditions of optimality. Let $\left\{x_{k}\right\}$ be a sequence converging to $x_{*}$, and $d_{k}$ be a stationary point of the osculating quadratic problem (17.30). In this last problem, suppose that the matrix $M_{k}$ over-estimates $L_{*}^{r}=L_{*}+r A_{*}^{\top} A_{*}$ in the sense that

$$
\begin{equation*}
d_{k}^{\top}\left(M_{k}-L_{*}^{r}\right) d_{k} \geq o\left(\left\|d_{k}\right\|^{2}\right) \tag{17.39}
\end{equation*}
$$

where $r \geq 0$ is such that $L_{*}^{r}$ is positive definite (such an $r$ always exists under the assumptions already stated, see exercise 16.1). Assume also that $d_{k} \rightarrow 0$, that $\omega<\frac{1}{2}$, and that $\sigma_{k} \geq\left\|\lambda_{k}^{\mathrm{QP}}-\mu_{k}\right\|_{D}$. Then there exists $\varepsilon>0$ such that, if $\left\|\mu_{k}-\lambda_{*}\right\| \leq \varepsilon$ and $0 \leq \sigma_{k} \leq \varepsilon$, we have for large enough $k$

$$
\ell_{\mu_{k}, \sigma_{k}}\left(x_{k}+d_{k}\right) \leq \ell_{\mu_{k}, \sigma_{k}}\left(x_{k}\right)+\omega \ell_{\mu_{k}, \sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)
$$

Proof. The following expansions are easily obtained:

$$
\begin{gathered}
f\left(x_{k}+d_{k}\right)=f_{k}+\nabla f_{k}^{\top} d_{k}+\frac{1}{2} d_{k}^{\top} \nabla^{2} f\left(x_{*}\right) d_{k}+o\left(\left\|d_{k}\right\|^{2}\right) \\
c\left(x_{k}+d_{k}\right)=\frac{1}{2} c^{\prime \prime}\left(x_{*}\right) \cdot d_{k}^{2}+o\left(\left\|d_{k}\right\|^{2}\right)
\end{gathered}
$$

We can then write

$$
\begin{aligned}
& \ell_{\mu_{k}, \sigma_{k}}\left(x_{k}+d_{k}\right)-\ell_{\mu_{k}, \sigma_{k}}\left(x_{k}\right)-\omega \ell_{\mu_{k}, \sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right) \\
&= \nabla f_{k}^{\top} d_{k}+\frac{1}{2} d_{k}^{\top} \nabla^{2} f\left(x_{*}\right) d_{k}+\frac{1}{2} \mu_{k}^{\top} c^{\prime \prime}\left(x_{*}\right) \cdot d_{k}^{2}-\mu_{k}^{\top} c_{k}-\sigma_{k}\left\|c_{k}\right\|_{P} \\
&-\omega \ell_{\mu_{k}, \sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)+O\left(\sigma_{k}\left\|d_{k}\right\|^{2}\right)+o\left(\left\|d_{k}\right\|^{2}\right) \\
&=(1-\omega) \ell_{\mu_{k}, \sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)+\frac{1}{2} d_{k}^{\top} L_{*} d_{k} \\
&+O\left(\left(\left\|\mu_{k}-\lambda_{*}\right\|_{D}+\sigma_{k}\right)\left\|d_{k}\right\|^{2}\right)+o\left(\left\|d_{k}\right\|^{2}\right) \\
& \leq(1-\omega) \ell_{\mu_{k}, \sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)+\frac{1}{2} d_{k}^{\top} L_{*}^{r} d_{k}+C_{1} \varepsilon\left\|d_{k}\right\|^{2}+o\left(\left\|d_{k}\right\|^{2}\right) \\
& \leq\left(\frac{1}{2}-\omega\right) \ell_{\mu_{k}, \sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)-\frac{1}{2} d_{k}^{\top}\left(M_{k}-L_{*}^{r}\right) d_{k}+C_{1} \varepsilon\left\|d_{k}\right\|^{2}+o\left(\left\|d_{k}\right\|^{2}\right) \\
& \leq-C_{2}\left(\frac{1}{2}-\omega\right)\left\|d_{k}\right\|^{2}+C_{1} \varepsilon\left\|d_{k}\right\|^{2}+o\left(\left\|d_{k}\right\|^{2}\right) \\
& \leq 0
\end{aligned}
$$

if $k$ is large enough and $\varepsilon>0$ is small enough. We have used the uniform positive definiteness of $M_{k}$, which comes from the positive definiteness of $L_{*}^{r}$ and from (17.39).

We refer the reader to the original paper [37] and to [146, 10] for examples of use of the nondifferentiable augmented Lagrangian in implementable algorithms.

### 17.4 The Hanging Chain Project IV

This is the fourth session dealing with the problem of finding the static equilibrium of chain made of rigid bars that stays above a given tilted floor. The
problem was introduced in $\S 13.8$ and developed in $\S \S 14.7$ and 15.4. We now consider the implementation of the globalization technique presented in this chapter. This will provide more robustness to the SQP solver and will give it a tendency to avoid the stationary points that are not local minima.

We propose to use the merit function (17.1) in which $\|\cdot\|_{P}$ is the $\ell_{1}$ norm $\|v\|_{1}:=\sum_{i=1}^{m}\left|v_{i}\right|:$

$$
\begin{equation*}
\Theta_{\sigma}(x)=f(x)+\sigma\left\|c(x)^{\#}\right\|_{1} \tag{17.40}
\end{equation*}
$$

This norm satisfies the assumption (17.5) required by proposition 17.1 (see exercise 17.1). The dual norm of the $\ell_{1}$ norm is the $\ell_{\infty}$ norm $\|w\|_{\infty}:=$ $\max _{1 \leq i \leq m}\left|w_{i}\right|$ (see exercise 16.5).

We assume that the osculating quadratic program has the form (17.2), with a matrix $M_{k}$ that is symmetric positive definite. This property of $M_{k}$ is important in order to get a primal solution $d_{k}$ to (17.2) that is a descent direction of the exact merit function $\Theta_{\sigma}$ defined by (17.40) (see proposition 17.1). Since the Hessian of the Lagrangian $L_{k}:=\nabla_{x x}^{2} \ell\left(x_{k}, \lambda_{k}\right)$ is not necessarily positive definite, we propose to take for $M_{k}$ a modification of $L_{k}$ obtained by adding to it a small positive diagonal matrix (using, for example, a modified Cholesky factorization $[154,201])$. Using a positive definite quasiNewton approximation to $L_{k}$ is another possibility that will be examined in chapter 18.

## Modifications to Bring to the sqp Function

It is interesting to keep the possibility of using the algorithms defined in the previous sessions by introducing flags. In our code, we use options.imode (1:2), which has the following meanings:

- imode (1): $0\left(M_{k}\right.$ is a quasi-Newton approximation to $\left.L_{k}\right), 1\left(M_{k}=L_{k}\right)$, $2\left(M_{k}=L_{k}+E_{k}\right.$, where $E_{k}$ is a small positive diagonal matrix that makes $M_{k}$ positive definite),
- imode(2): 0 (with line-search), 1 (with unit stepsize).

If we compare the local SQP algorithm on page 257 , implemented in the previous sessions, and the version with line-search on page 292, we see that we essentially have to add the steps 3,4 , and 5 of the latter algorithm to the sqp function.

- The determination of the penalty parameter $\sigma_{k}$ in step 3 can be done by the update rule of page 295. At the first iteration, we take $\sigma_{1}=\left\|\lambda_{1}^{\mathrm{QP}}\right\|_{D}+\bar{\sigma}$ and set the constant $\bar{\sigma}$ to $\max \left(\sqrt{\mathrm{eps}},\left\|\lambda_{1}^{\mathrm{QP}}\right\|_{D} / 100\right)$.
- The determination of a stepsize $\alpha_{k}$ along $d_{k}$ in step 4 can be done like in the backtracking line-search of page 296 , with $\beta=0.1$ and $\alpha_{k, i+1}$ determined by interpolation, i.e., as the minimizer of the quadratic function $\alpha \mapsto \xi(\alpha)$ satisfying $\xi(0)=\Theta_{\sigma_{k}}\left(x_{k}\right), \xi^{\prime}(0)=\Delta_{k}$, and $\xi\left(\alpha_{k, i}\right)=$ $\Theta_{\sigma_{k}}\left(x_{k}+\alpha_{k, i} d_{k}\right)$.
- We set the new multiplier $\lambda_{k+1}$ by (17.4).

It is better not to limit the number of stepsize trials in the line-search, since this number, which is most often 1 , can be large at some difficult iteration. However, the line-search algorithm may cycle when there is an error in the simulator or when rounding errors occur at the end of a minimization. Therefore, some arrangements have to be implemented to prevent this cycling. In our code, the line-search is interrupted when the norm of the step $\alpha_{k, i}\left\|d_{k}\right\|_{\infty}$ to get improvement in the merit function becomes smaller than a prescribed value options.dxmin given on entry in the solver.

It is important to take care over the output printed by the code, since it provides meaningful information on the course of the optimization. Here is the text, in connection with the line-search, that our code prints at each iteration.

```
iter 11, simul 14, merit -1.47914e+00, slope -7.59338e-02
    Armijo's line-search
    1.0000e+00 8.47489e-01 8.47489e-01
    1.0000e-01 1.49986e-03 1.49986e-02
    4.1753e-02 -1.60114e-03 -3.83479e-02
```

The value of $\Delta_{k}$ defined by (17.7), which approximates $\Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)$, is given after the keyword slope, and should always be negative. Each line of the table below the phrase "Armijo's line-search" corresponds to a stepsize trial: $\alpha_{k, i}$ is in the first column, $\Theta_{\sigma_{k}}\left(x_{k}+\alpha_{k, i} d_{k}\right)-\Theta_{\sigma_{k}}\left(x_{k}\right)$ in the second, and $\left(\Theta_{\sigma_{k}}\left(x_{k}+\alpha_{k, i} d_{k}\right)-\Theta_{\sigma_{k}}\left(x_{k}\right)\right) / \alpha_{k, i}$ in the last one. We see in the first column that the unit stepsize $\alpha_{k, 1}=1$ is tried first and that it is determined next by interpolation with the safeguard $\beta=0.1$. The last column is useful to detect a possible inconsistency in the simulator (or in the sqp function). If $d_{k}$ is not a descent direction of the merit function $\Theta_{\sigma_{k}}$ (it should be a descent direction if $M_{k}$ is positive definite and if nothing is wrong in the simulator and in the sqp function, see proposition 17.1), there is a large number of stepsize trials $\alpha_{k, i}$ tending to zero. Then, the value in the last column should tend to $\Delta_{k}$ (this is actually certainly correct if there is no inequality constraint, since then $\Delta_{k}=\Theta_{\sigma_{k}}^{\prime}\left(x_{k} ; d_{k}\right)$, see the comment after proposition 17.1).
Question: Tell why the last value in the third column of the table after the phrase "Armijo's line-search" above is often approximately half that of $\Delta_{k}$ (like here: $3.83479 / 7.59338 \simeq 0.505$ ).

## Experimenting with the SQP Algorithm

The first observation is good news: line-search really helps to force convergence. For example, test case 1d (page 249), which diverges without linesearch, now converges to the global minimum. Figure 17.5 shows the result with the usual convention: the thin solid bars represent the initial position of the chain, the dashed bars correspond to the intermediate positions, and the


Fig. 17.5. Test case 1 d with line-search
bold solid bars are those of the final optimal position. For clarity, we have not represented all the intermediate positions of the 10 iterations required to get convergence, but 1 out of 2 .

The second observation is that line-search helps the SQP algorithm to avoid stationary points that are not local minima. For example if we apply the present algorithm with line-search to test case 1b (page 249), the generated sequence now converges to the global minimum of the problem, not to the global maximum as before. The left picture in figure 17.6 shows the result (1


Fig. 17.6. Test cases 1 b (left) and 1c (right) with line-search
iteration out of 3 ). The same phenomenon occurs with test case 1c (page 249),
whose convergence to the global minimum is shown in the right hand side picture of figure 17.6.

A third observation: the convergence is smoother with line-search. This is not a very precise concept, but we mean by this that the behavior of the generated sequence is less erratic. Consider for example test case 1f (page 269). The result is shown in figure 17.7. If we compare with figure 15.3, we see that


Fig. 17.7. Test case 1f with line-search
the second iterate is now closer the the initial one: the stepsize is actually less than $1\left(\alpha_{1}=0.1\right)$ only at the first iteration. This additional function evaluation is beneficial since the total number of function evaluations is less than the one without line-search (10 instead of 11, not a major improvement, admittedly).

## Notes

The use of the exact penalty function (17.1) to globalize the SQP algorithm was proposed by Pshenichnyj (see for example [302]), Han [185; 1977] (with the $\ell_{1}$ norm), and others. The TSQP algorithm described in $\S 17.2$ is taken from [75; 2003]. Another way of dealing with nonconvex problems is to modify the Hessian of the Lagrangian, using a modified Cholesky factorization (see for example [133] and the references therein).

The "effect" described in § 17.3 was discovered by Maratos [247; 1978] and counter-example 17.6 is adapted from [73]. Second-order correction strategies were proposed by Boggs, Tolle, and Wang [36], Coleman and Conn [82], Fletcher [127], Gabay [138], Mayne and Polak [250]. The use of the nondifferentiable augmented Lagrangian was proposed by Bonnans [37]. Note that Fletcher's exact penalty function (16.21) also accepts the unit stepsize asymptotically, but it involves first derivatives, so that its use may lead to expensive algorithms if a number of different stepsizes are required during the line-search or to algorithmic remedies for avoiding expensive operations;
see $[299,33,34]$. Other approaches include the "watchdog" technique [73] and the nonmonotone line-search [281, 46].

To conclude this chapter let us briefly mention and/or review other contributions dealing with the use of second derivatives within SQP, techniques for solving the QP, and algorithmic modifications for tackling large-scale problems: Betts and Frank [29] add a positive multiple of the identity matrix to the full Hessian of the Lagrangian when the factorization of the KKT matrix reveals nonpositive definiteness of the reduced Hessian of the Lagrangian; Bonnans and Launay [45]; Murray and Prieto [270]; Gill, Murray, and Saunders [155]; Leibfritz and Sachs [225]; Facchinei and Lucidi [121]; Boggs, Kearsley, and Tolle [32, 31] propose solving the QP by an interior point method that can be prematurely halted by a pseudo-trust-region constraint, although their method uses line-search for its globalization; Sargent and Ding [321] also use an interior point method to solve the QP inexactly within a line-search approach, but discard the Hessian of the Lagrangian if it fails to yield a descent direction of the merit function; Byrd, Gilbert, and Nocedal [65] combine SQP with an interior point approach on the nonlinear problem and use trust regions for the globalization.

## Exercises

17.1. Norm assumptions. Let $\|\cdot\|$ be an arbitrary norm on $\mathbb{R}^{m}$ and consider the following properties (the operators $|\cdot|$ and $(\cdot)^{+}$act componentwise; the statements are valid for all $u$ and $v \in \mathbb{R}^{m}$ when this makes sense):
(i) $\||u|\|=\|u\|$;
(ii) $|u| \leq|v| \Longrightarrow\|u\| \leq\|v\|$;
(iii) $u_{i}=v_{i}$ or $0 \Longrightarrow\|u\| \leq\|v\|$;
(iv) $0 \leq u \leq v \quad \Longrightarrow \quad\|u\| \leq\|v\|$;
(v) $u \leq v \Longrightarrow\left\|u^{+}\right\| \leq\left\|v^{+}\right\|$;
(vi) $v \mapsto\left\|v^{+}\right\|$is convex.

Show that $(i) \Leftrightarrow(i i) \Rightarrow(i i i) \Rightarrow(i v) \Leftrightarrow(v) \Leftrightarrow(v i)$, but that none of the other implications holds in general. Show that (vi) may not hold for an arbitrary norm.

Remark: These implications show that assumptions (16.15) and (17.5) on the norm $\|\cdot\|_{P}$ are satisfied with the $\ell_{p}$ norms, $1 \leq p \leq \infty$, since $\ell_{p}$ norms satisfy $(i)$. They also show that (16.15) is more restrictive than (17.5).
17.2. On the directional derivative of $\Theta_{\sigma}$. Find a one-dimensional example, in which $\Theta_{\sigma}^{\prime}(x ; d)<\nabla f(x)^{\top} d-\sigma\left\|c(x)^{\#}\right\|_{P}$, where $d$ is the solution to the osculating quadratic problem (17.2) (hence the inequality in (17.6) may be strict).
[Hint: Equality holds if $I=\emptyset$. .]
17.3. Descent direction for the exact penalization of the Lagrangian. Consider the exact penalty function $\Theta_{\mu, \sigma}: \mathbb{R}^{n} \rightarrow \mathbb{R}$ defined for $\mu \in \mathbb{R}^{m}$ and $\sigma>0$ by

$$
\Theta_{\mu, \sigma}(x):=f(x)+\mu^{\top} c(x)^{\#}+\sigma\left\|c(x)^{\#}\right\|_{P},
$$

where the norm $\|\cdot\|_{P}$ satisfies (17.5) (see also exercise 16.9). Let $\left(d_{k}, \lambda_{k}^{\mathrm{QP}}\right)$ satisfy the optimality conditions (17.3). Show that $d_{k}$ is a descent direction of $\Theta_{\mu, \sigma}$ at $x_{k}$, provided $x_{k}$ is not a stationary point of $\left(P_{E I}\right), M_{k}$ is positive definite, $\sigma \geq\left\|\lambda_{k}^{Q P}-\mu\right\|_{D}$, and $\mu_{I} \geq 0$.
17.4. Admissibility of the unit stepsize for Newton's method. Consider the problem of finding a root $x_{*}$ of the equation $F(x)=0$, where $F: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ is a smooth function. Newton's method consists in updating $x$ by $x_{+}=x+d$, where $d$ solves $F^{\prime}(x) d=-F(x)$ (see $\S 13.7$ ). Let $\|\cdot\|$ be an arbitrary norm and consider $\varphi(x)=\|F(x)\|$ as a merit function for this problem. Suppose that $F^{\prime}\left(x_{*}\right)$ is nonsingular. Show that, for any constant $\left.\omega \in\right] 0,1[$, there is a neighborhood $V$ of $x_{*}$, such that if $x \in V, \varphi(x+d) \leq \varphi(x)+\omega \varphi^{\prime}(x ; d)$.

## References

1. I. Adler, M.G.C. Resende, and G. Veiga. An implementation of Karmarkar's algorithm for linear programming. Math. Programming, 44:297-335, 1989.
2. M. Al-Baali. Descent property and global convergence of the Fletcher-Reeves methods with inexact line search. IMA Journal of Numerical Analysis, 5:121124, 1985.
3. W. Alt. The Lagrange-Newton method for infinite-dimensional optimization problems. Numerical Functional Analysis and Optimization, 11:201-224, 1990.
4. W. Alt. Sequential quadratic programming in Banach spaces. In W. Oettli and D. Pallaschke, editors, Advances in Optimization, number 382 in Lecture Notes in Economics and Mathematical Systems, pages 281-301. SpringerVerlag, 1992.
5. W. Alt. Semi-local convergence of the Lagrange-Newton method with application to optimal control. In R. Durier and Ch. Michelot, editors, Recent Developments in Optimization, number 429 in Lecture Notes in Economics and Mathematical Systems, pages 1-16. Springer-Verlag, 1995.
6. W. Alt and K. Malanowski. The Lagrange-Newton method for nonlinear optimal control problems. Computational Optimization and Applications, 2:77100, 1991.
7. M. Anitescu. Degenerate nonlinear programming with a quadratic growth condition. SIAM Journal on Optimization, 10:1116-1135, 2000.
8. M. Anitescu. On the rate of convergence of sequential quadratic programming with nondifferentiable exact penalty function in the presence of constraint degeneracy. Mathematical Programming, 92:359-386, 2002.
9. K.M. Anstreicher and M.H. Wright. A note on the augmented Hessian when the reduced Hessian is semidefinite. SIAM Journal on Optimization, 11:243253, 2000.
10. P. Armand and J.Ch. Gilbert. A piecewise line-search technique for maintaining the positive definiteness of the updated matrices in the SQP method. Computational Optimization and Applications, 16:121-158, 2000.
11. P. Armand, J.Ch. Gilbert, and S. Jan-Jégou. A BFGS-IP algorithm for solving strongly convex optimization problems with feasibility enforced by an exact penalty approach. Mathematical Programming, 92:393-424, 2002.
12. L. Armijo. Minimization of functions having Lipschitz continuous first partial derivatives. Pacific Journal of Mathematics, 16:1-3, 1966.
13. K. Arrow, L. Hurwicz, and H. Uzawa. Studies in Nonlinear Programming. Stanford University Press, Stanford, CA, 1958.
14. K.J. Arrow and R.M. Solow. Gradient methods for constrained maxima with weakened assumptions. In K.J. Arrow, L. Hurwicz, and H. Uzawa, editors,

Studies in Linear and Nonlinear Programming. Stanford University Press, Standford, Calif., 1958.
15. A. Auslender. Numerical methods for nondifferentiable convex optimization. Mathematical Programming Study, 30:102-126, 1987.
16. A. Auslender and M. Teboulle. Lagrangian duality and related multiplier methods for variational inequality problems. SIAM Journal on Optimization, 10:1097-1115, 2000.
17. A. Auslender, M. Teboulle, and S. Ben-Tiba. Interior proximal and multiplier methods based on second order homogeneous functionals. Mathematics of Operations Research, 24:645-668, 1999.
18. F. Babonneau, C. Beltran, A. Haurie, C. Tadonki, and J.-P. Vial. ProximalACCPM: a versatile oracle based optimization method. To appear in Computational Management Science, 2006.
19. L. Bacaud, C. Lemaréchal, A. Renaud, and C. Sagastizábal. Bundle methods in stochastic optimal power management: a disaggregated approach using preconditioners. Computational Optimization and Applications, 20(3):227-244, 2001.
20. R.E. Bank, B.D. Welfert, and H. Yserentant. A class of iterative methods for solving saddle point problems. Numerische Mathematik, 56:645-666, 1990.
21. A. Belloni, A. Diniz, M.E. Maceira, and C. Sagastizábal. Bundle relaxation and primal recovery in Unit Commitment problems. The Brazilian case. Annals of Operations Research, 120:21-44, 2003.
22. A. Belloni and C. Sagastizábal. Dynamic bundle methods: Application to combinatorial optimization. Technical report, Optimization on line, 2004. http://www.optimization-online.org/DB_HTML/2004/08/925.html.
23. A. Ben-Tal and A. Nemirovski. Lectures on Modern Convex Optimization Analysis, Algorithms, and Engineering Applications. MPS/SIAM Series on Optimization 2. SIAM, 2001.
24. M.W. Berry, M.T. Health, I. Kaneko, M. Lawo, R.J. Plemmons, and R.C. Ward. An algorithm to compute a sparse basis of the null space. Numerische Mathematik, 47:483-504, 1985.
25. D.P. Bertsekas. Multiplier methods: a survey. Automatica, 12:133-145, 1976.
26. D.P. Bertsekas. Constrained Optimization and Lagrange Multiplier Methods. Academic Press, 1982.
27. D.P. Bertsekas. Nonlinear Programming. Athena Scientific, 1995. Second edition, 1999.
28. J.T. Betts. Practical Methods for Optimal Control Using Nonlinear Programming. SIAM, 2001.
29. J.T. Betts and P.D. Frank. A sparse nonlinear optimization algorithm. Journal of Optimization Theory and Applications, 3:519-541, 1994.
30. L.T. Biegler, J. Nocedal, and C. Schmid. A reduced Hessian method for largescale constrained optimization. SIAM Journal on Optimization, 5:314-347, 1995.
31. P.T. Boggs, A.J. Kearsley, and J.W. Tolle. A global convergence analysis of an algorithm for large-scale nonlinear optimization problems. SIAM Journal on Optimization, 9:833-862, 1999.
32. P.T. Boggs, A.J. Kearsley, and J.W. Tolle. A practical algorithm for general large scale nonlinear optimization problems. SIAM Journal on Optimization, 9:755-778, 1999.
33. P.T. Boggs and J.W. Tolle. A family of descent functions for constrained optimization. SIAM Journal on Numerical Analysis, 21:1146-1161, 1984.
34. P.T. Boggs and J.W. Tolle. A strategy for global convergence in a sequential quadratic programming algorithm. SIAM Journal on Numerical Analysis, 26:600-623, 1989.
35. P.T. Boggs and J.W. Tolle. Sequential quadratic programming. In Acta Numerica 1995, pages 1-51. Cambridge University Press, 1995.
36. P.T. Boggs, J.W. Tolle, and P. Wang. On the local convergence of quasiNewton methods for constrained optimization. SIAM Journal on Control and Optimization, 20:161-171, 1982.
37. J.F. Bonnans. Asymptotic admissibility of the unit stepsize in exact penalty methods. SIAM Journal on Control and Optimization, 27:631-641, 1989.
38. J.F. Bonnans. Local study of Newton type algorithms for constrained problems. In S. Dolecki, editor, Optimization, number 1405 in Lecture Notes in Mathematics, pages 13-24. Springer-Verlag, 1989.
39. J.F. Bonnans. Théorie de la pénalisation exacte. Modélisation Mathématique et Analyse Numérique, 24:197-210, 1990.
40. J.F. Bonnans. Local analysis of Newton-type methods for variational inequalities and nonlinear programming. Applied Mathematics and Optimization, 29:161-186, 1994.
41. J.F. Bonnans. Exact penalization with a small nonsmooth term. Revista de Matemáticas Aplicadas, 17:37-45, 1996.
42. J.F. Bonnans, J.Ch. Gilbert, C. Lemaréchal, and C. Sagastizábal. A family of variable metric proximal methods. Mathematical Programming, 68:15-47, 1995.
43. J.F. Bonnans, J.Ch. Gilbert, C. Lemaréchal, and C. Sagastizábal. Optimisation Numérique - Aspects théoriques et pratiques. Number 27 in Mathématiques et Applications. Springer Verlag, Berlin, 1997.
44. J.F. Bonnans and C.C. Gonzaga. Convergence of interior point algorithms for the monotone linear complementarity problem. Mathematics of Operations Research, 21:1-25, 1996.
45. J.F. Bonnans and G. Launay. Sequential quadratic programming with penalization of the displacement. SIAM Journal on Optimization, 5:792-812, 1995.
46. J.F. Bonnans, E.R. Panier, A.L. Tits, and J.L. Zhou. Avoiding the Maratos effect by means of a nonmonotone line search II: Inequality constrained problems - Feasible iterates. SIAM Journal on Numerical Analysis, 29:1187-1202, 1992.
47. J.F. Bonnans, C. Pola, and R. Rebaï. Perturbed path following interior point algorithms. Optimization Methods and Software, 11-12:183-210, 1999.
48. J.F. Bonnans and F.A. Potra. Infeasible path following algorithms for linear complementarity problems. Mathematics of Operations Research, 22:378-407, 1997.
49. J.F. Bonnans and A. Shapiro. Optimization problems with perturbations - A guided tour. SIAM Review, 40:202-227, 1998.
50. J.F. Bonnans and A. Shapiro. Perturbation Analysis of Optimization Problems. Springer Verlag, New York, 2000.
51. W. Boothby. An Introduction to Differentiable Manifolds and Differential Geometry. Academic Press, New York, 1975.
52. J. Borwein and A.S. Lewis. Convex Analysis and Nonlinear Optimization. Springer Verlag, New York, 2000.
53. J.H. Bramble and J.E. Pasciak. A preconditioning technique for indefinite systems resulting from mixed approximations of elliptic problems. Mathematics of Computation, 50:1-7, 1988.
54. J.H. Bramble, J.E. Pasciak, and A.T. Vassilev. Analysis of the Uzawa algorithm for saddle point problems. SIAM Journal on Numerical Analysis, 34:1072-1092, 1997.
55. U. Brännlund. On relaxation methods for nonsmooth convex optimization. PhD thesis, Royal Institute of Technology - Stockholm, 1993.
56. J.R. Bunch and L. Kaufman. Some stable methods for calculating inertia and solving symmetric linear systems. Mathematics of Computation, 31:163-179, 1977.
57. R.S. Burachik, A.N. Iusem, and B.F. Svaiter. Enlargement of monotone operators with applications to variational inequalities. Set-Valued Anal., 5(2):159180, 1997.
58. R.S. Burachik, C. Sagastizábal, and S. Scheinberg de Makler. An inexact method of partial inverses and a parallel bundle method. Optimization Methods and Software, 21(3):385-400, 2006.
59. R.S. Burachik, C. Sagastizábal, and B. F. Svaiter. Bundle methods for maximal monotone operators. In R. Tichatschke and M. Théra, editors, Ill-posed Variational Problems and Regularization Techniques, number 477 in Lecture Notes in Economics and Mathematical Systems, pages 49-64. Springer-Verlag Berlin Heidelberg, 1999.
60. J.V. Burke. An exact penalization viewpoint of constrained optimization. SIAM Journal on Control and Optimization, 29:968-998, 1991.
61. J.V. Burke and S.-P. Han. A robust sequential quadratic programming method. Mathematical Programming, 43:277-303, 1989.
62. J.D. Buys. Dual algorithms for constrained optimization. PhD thesis, Rijksuniversiteit te Leiden, Leiden, The Netherlands, 1972.
63. R.H. Byrd. An example of irregular convergence in some constrained optimization methods that use the projected Hessian. Mathematical Programming, 32:232-237, 1985.
64. R.H. Byrd. On the convergence of constrained optimization methods with accurate Hessian information on a subspace. SIAM Journal on Numerical Analysis, 27:141-153, 1990.
65. R.H. Byrd, J.Ch. Gilbert, and J. Nocedal. A trust region method based on interior point techniques for nonlinear programming. Mathematical Programming, 89:149-185, 2000.
66. R.H. Byrd and J. Nocedal. A tool for the analysis of quasi-Newton methods with application to unconstrained minimization. SIAM Journal on Numerical Analysis, 26:727-739, 1989.
67. R.H. Byrd and J. Nocedal. An analysis of reduced Hessian methods for constrained optimization. Mathematical Programming, 49:285-323, 1991.
68. R.H. Byrd and R.B. Schnabel. Continuity of the null space basis and constrained optimization. Mathematical Programming, 35:32-41, 1986.
69. R.H. Byrd, R.A. Tapia, and Y. Zhang. An SQP augmented Lagrangian BFGS algorithm for constrained optimization. SIAM Journal on Optimization, 2:210-241, 1992.
70. A. Cauchy. Méthode générale pour la résolution des systèmes d'équations simultanées. C. R. Acad. Sci. Paris, 25:535-538, 1847.
71. J. Céa. Optimisation: Théorie et Algorithmes. Dunod, Paris, 1971.
72. Y. Chabrillac and J.-P. Crouzeix. Definiteness and semidefiniteness of quadratic forms revisited. Linear Algebra and its Applications, 63:283-292, 1984.
73. R.M. Chamberlain, C. Lemaréchal, H.C. Pedersen, and M.J.D. Powell. The watchdog technique for forcing convergence in algorithms for constrained optimization. Mathematical Programming Study, 16:1-17, 1982.
74. C. Charalambous. A lower bound for the controlling parameters of the exact penalty functions. Mathematical Programming, 15:278-290, 1978.
75. L. Chauvier, A. Fuduli, and J.Ch. Gilbert. A truncated SQP algorithm for solving nonconvex equality constrained optimization problems. In G. Di Pillo and A. Murli, editors, High Performance Algorithms and Software for Nonlinear Optimization, pages 146-173. Kluwer Academic Publishers B.V., 2003.
76. G. Chen and M. Teboulle. A proximal-based decomposition method for convex minimization problems. Math. Programming, 64(1, Ser. A):81-101, 1994.
77. X. Chen and M. Fukushima. Proximal quasi-Newton methods for nondifferentiable convex optimization. Math. Program., 85(2, Ser. A):313-334, 1999.
78. E. Cheney and A. Goldstein. Newton's method for convex programming and Tchebycheff approximations. Numerische Mathematik, 1:253-268, 1959.
79. P.G. Ciarlet. Introduction à l'Analyse Numérique Matricielle et à l'Optimisation (second edition). Masson, Paris, 1988.
80. F.H. Clarke. Optimization and Nonsmooth Analysis. John Wiley \& Sons, New York; reprinted by SIAM, 1983.
81. T.F. Coleman and A.R. Conn. Nonlinear programming via an exact penalty function: asymptotic analysis. Mathematical Programming, 24:123-136, 1982.
82. T.F. Coleman and A.R. Conn. Nonlinear programming via an exact penalty function: global analysis. Mathematical Programming, 24:137-161, 1982.
83. T.F. Coleman and D.C. Sorensen. A note on the computation of an orthonormal basis for the null space of a matrix. Mathematical Programming, 29:234242, 1984.
84. L. Conlon. Differentiable Manifolds - A first Course. Birkhauser, Boston, 1993.
85. A.R. Conn, N.I.M. Gould, and Ph.L. Toint. LANCELOT: A Fortran Package for Large-Scale Nonlinear Optimization (Release A). Number 17 in Computational Mathematics. Springer Verlag, Berlin, 1992.
86. A.R. Conn, N.I.M. Gould, and Ph.L. Toint. Trust-Region Methods. MPS/SIAM Series on Optimization. MPS/SIAM, Philadelphia, 2000.
87. D. Coppersmith and S. Winograd. On the asymptotic complexity of matrix multiplications. SIAM J. Computation, 11:472-492, 1982.
88. G. Corliss and A. Griewank, editors. Automatic Differentiation of Algorithms: Theory, Implementation, and Application. Proceedings in Applied Mathematics 53. SIAM, Philadelphia, 1991.
89. R. Correa and C. Lemaréchal. Convergence of some algorithms for convex minimization. Mathematical Programming, 62:261-275, 1993.
90. R.W. Cottle. Manifestations of the Schur complement. Linear Algebra and its Applications, 8:189-211, 1974.
91. R.W. Cottle, J.S. Pang, and R.E. Stone. The linear complementarity problem. Academic Press, New York, 1992.
92. M. Cui. A sufficient condition for the convergence of the inexact Uzawa algorithm for saddle point problems. Journal of Computational and Applied Mathematics, 139:189-196, 2002.
93. J.-C. Culioli and G. Cohen. Decomposition/coordination algorithms in stochastic optimization. SIAM J. Control Optim., 28(6):1372-1403, 1990.
94. G.B. Dantzig. Linear programming and extensions. Princeton University Press, Princeton, N.J., 1963.
95. G.B. Dantzig and P. Wolfe. The decomposition algorithm for linear programming. Econometrica, 29(4):767-778, 1961.
96. W.C. Davidon. Variable metric methods for minimization. AEC Research and Development Report ANL-5990, Argonne National Laboratory, Argonne, Illinois, 1959.
97. W.C. Davidon. Variable metric method for optimization. SIAM Journal on Optimization, 1:1-17, 1991.
98. F. Delbos and J.Ch. Gilbert. Global linear convergence of an augmented Lagrangian algorithm for solving convex quadratic optimization problems. Journal of Convex Analysis, 12:45-69, 2005.
99. F. Delbos, J.Ch. Gilbert, R. Glowinski, and D. Sinoquet. Constrained optimization in seismic reflection tomography: a Gauss-Newton augmented Lagrangian approach. Geophysical Journal International, 164:670-684, 2006.
100. F. Delprat-Jannaud and P. Lailly. What information on the Earth model do reflection travel times hold? Journal of Geophysical Research, 97:19827-19844, 1992.
101. F. Delprat-Jannaud and P. Lailly. Ill-posed and well-posed formulations of the reflection travel time tomography problem. Journal of Geophysical Research, 98:6589-6605, 1993.
102. R.S. Dembo and T. Steihaug. Truncated-Newton algorithms for large-scale unconstrained optimization. Mathematical Programming, 26:190-212, 1983.
103. D. den Hertog. Interior-point approach to linear, quadratic and convex programming. Kluwer Academic Publishers, Boston, 1994.
104. J.E. Dennis and J.J. Moré. A characterization of superlinear convergence and its application to quasi-Newton methods. Mathematics of Computation, 28:549-560, 1974.
105. J.E. Dennis and J.J. Moré. Quasi-Newton methods, motivation and theory. SIAM Review, 19:46-89, 1977.
106. J.E. Dennis and R.B. Schnabel. A new derivation of symmetric positive definite secant updates. In Nonlinear Programming 4, pages 167-199. Academic Press, 1981.
107. J.E. Dennis and R.B. Schnabel. Numerical Methods for Unconstrained Optimization and Nonlinear Equations. Prentice-Hall, Englewood Cliffs, 1983.
108. P. Deuflhard. Newton Methods for Nonlinear Problems - Affine Invariance and Adaptative Algorithms. Number 35 in Computational Mathematics. Springer, Berlin, 2004.
109. G. Di Pillo and L. Grippo. A new class of augmented Lagrangians in nonlinear programming. SIAM Journal on Control and Optimization, 17:618-628, 1979.
110. G. Di Pillo and S. Lucidi. On exact augmented Lagrangian functions in nonlinear programming. In G. Di Pillo and F. Giannessi, editors, Nonlinear Optimization and Applications, pages 85-100. Plenum Press, New York, 1996.
111. G. Di Pillo and S. Lucidi. An augmented Lagrangian function with improved exactness properties. SIAM Journal on Optimization, 12:376-406, 2001.
112. M.P. do Carmo. Riemannian Geometry. Birkhauser, Boston, 1993.
113. E.D. Dolan and J.J. Moré. Benchmarking optimization software with performance profiles. Mathematical Programming, 91:201-213, 2002.
114. I.S. Duff and J.K. Reid. MA27 - A set of Fortran subroutines for solving sparse symmetric sets of linear equations. Technical Report AERE R10533, HMSO, London, 1982.
115. I.S. Duff and J.K. Reid. The multifrontal solution of indefinite sparse symmetric linear systems. ACM Transactions on Mathematical Software, 9:301-325, 1983.
116. I.S. Duff and J.K. Reid. Exploiting zeros on the diagonal in the direct solution of indefinite sparse symmetric linear systems. ACM Transactions on Mathematical Software, 22:227-257, 1996.
117. I. Ekeland and R. Temam. Analyse convexe et problèmes variationnels. Dunod-Gauthier Villars, Paris, 1974.
118. H.C. Elman and G.H. Golub. Inexact and preconditioned Uzawa algorithms for saddle point problems. SIAM Journal on Numerical Analysis, 31:16451661, 1994.
119. I.I. Eremin. The penalty method in convex programming. Soviet Mathematics Doklady, 8:459-462, 1966.
120. F. Facchinei. Exact penalty functions and Lagrange multipliers. Optimization, 22:579-606, 1991.
121. F. Facchinei and S. Lucidi. Convergence to second-order stationary points in inequality constrained optimization. Mathematics of Operations Research, 23:746-766, 1998.
122. A.V. Fiacco and G.P. McCormick. Nonlinear Programming: sequential unconstrained minimization technique. J. Wiley, New York, 1968.
123. P. Finsler. Über das vorkommen definiter und semidefiniter formen und scharen quadratischer formen. Commentarii Mathematici Helvetica, 9:188192, 1937.
124. R. Fletcher. A class of methods for nonlinear programming with termination and convergence properties. In J. Abadie, editor, Integer and Nonlinear Programming. North-Holland, Amsterdam, 1970.
125. R. Fletcher. A FORTRAN subroutine for quadratic programming. Report R 6370, Atomic Energy Research Establishment, Harwell, England, 1970.
126. R. Fletcher. A model algorithm for composite nondifferentiable optimization problems. Mathematical Programming Study, 17:67-76, 1982.
127. R. Fletcher. Second order corrections for non-differentiable optimization. In D. Griffiths, editor, Numerical Analysis, pages 85-114. Springer-Verlag, 1982.
128. R. Fletcher. Practical Methods of Optimization (second edition). John Wiley \& Sons, Chichester, 1987.
129. R. Fletcher and T. Johnson. On the stability of null-space methods for KKT systems. SIAM Journal on Matrix Analysis and Applications, 18:938-958, 1997.
130. R. Fletcher and S. Leyffer. Nonlinear programming without a penalty function. Mathematical Programming, 91:239-269, 2002.
131. R. Fletcher and S. Leyffer. A bundle filter method for nonsmooth nonlinear optimization. Technical report, University of Dundee Numerical Analysis Report NA 195, December, 1999. http://www-unix.mcs.anl.gov/~leyffer/ papers/nsfilter.pdf.
132. R. Fletcher and M.J.D. Powell. A rapidly convergent descent method for minimization. The Computer Journal, 6:163-168, 1963.
133. A. Forsgren, P.E. Gill, and W. Murray. Computing modified Newton directions using a partial Cholesky factorization. SIAM Journal on Scientific Computing, 16:139-150, 1995.
134. M. Fortin and R. Glowinski. Méthodes de Lagrangien Augmenté - Applications à la Résolution Numérique de Problèmes aux Limites. Number 9 in Méthodes Mathématiques de l'Informatique. Dunod, Paris, 1982.
135. A. Frangioni. Generalized bundle methods. SIAM Journal on Optimization, 13(1):117-156, 2003.
136. M. Fukushima. A descent algorithm for nonsmooth convex optimization. Mathematical Programming, 30:163-175, 1984.
137. D. Gabay. Minimizing a differentiable function over a differential manifold. Journal of Optimization Theory and Applications, 37:177-219, 1982.
138. D. Gabay. Reduced quasi-Newton methods with feasibility improvement for nonlinearly constrained optimization. Mathematical Programming Study, 16:18-44, 1982.
139. Gauss. Theoria motus corporum coelestium. F. Perthes and I.H. Besser, Hamburg, 1809.
140. J. Gauvin. A necessary and sufficient regularity condition to have bounded multipliers in nonconvex programming. Mathematical Programming, 12:136138, 1977.
141. J. Gauvin. Théorie de la programmation mathématique non convexe. Les Publications CRM, Montréal, 1992.
142. J. Gauvin. Lecons de Programmation Mathématique. Éditions de l'École Polytechnique de Montréal, Montréal, 1995.
143. D.M. Gay, M.L. Overton, and M.H. Wright. A primal-dual interior method for nonconvex nonlinear programming. In Y.-X. Yuan, editor, Advances in Nonlinear Programming. Kluwer Academic Publishers, 1998.
144. J.Ch. Gilbert. Mise à jour de la métrique dans les méthodes de quasiNewton réduites en optimisation avec contraintes d'égalité. Modélisation Mathématique et Analyse Numérique, 22:251-288, 1988.
145. J.Ch. Gilbert. On the local and global convergence of a reduced quasi-Newton method. Optimization, 20:421-450, 1989.
146. J.Ch. Gilbert. Maintaining the positive definiteness of the matrices in reduced secant methods for equality constrained optimization. Mathematical Programming, 50:1-28, 1991.
147. J.Ch. Gilbert. Superlinear convergence of a reduced BFGS method with piecewise line-search and update criterion. Rapport de Recherche 2140, INRIA, BP 105, 78153 Le Chesnay, France, 1993.
148. J.Ch. Gilbert. On the realization of the Wolfe conditions in reduced quasiNewton methods for equality constrained optimization. SIAM Journal on Optimization, 7:780-813, 1997.
149. J.Ch. Gilbert. Piecewise line-search techniques for constrained minimization by quasi-Newton algorithms. In Y.-X. Yuan, editor, Advances in Nonlinear Programming, chapter 4, pages 73-103. Kluwer Academic Publishers, 1998.
150. J.Ch. Gilbert. Éléments d'Optimisation Différentiable - Théorie et Algorithmes. 2006. http://www-rocq.inria.fr/~gilbert/ensta/optim.html.
151. J.Ch. Gilbert, G. Le Vey, and J. Masse. La différentiation automatique de fonctions représentées par des programmes. Rapport de Recherche $\mathrm{n}^{\circ} 1557$, Inria, BP 105, F-78153 Le Chesnay, France, 1991.
152. J.Ch. Gilbert and C. Lemaréchal. Some numerical experiments with variablestorage quasi-Newton algorithms. Mathematical Programming, 45:407-435, 1989.
153. J.Ch. Gilbert and J. Nocedal. Global convergence properties of conjugate gradient methods for optimization. SIAM Journal on Optimization, 2:21-42, 1992.
154. P.E. Gill and W. Murray. Newton-type methods for unconstrained and linearly constrained optimization. Mathematical Programming, 7:311-350, 1974.
155. P.E. Gill, W. Murray, and M.A. Saunders. SNOPT: an SQP algorithm for large-scale constrained optimization. Numerical Analysis Report 96-2, Department of Mathematics, University of California, San Diego, La Jolla, CA, 1996.
156. P.E. Gill, W. Murray, and M.A. Saunders. SNOPT: an SQP algorithm for large-scale constrained optimization. SIAM Journal on Optimization, 12:9791006, 2002.
157. P.E. Gill, W. Murray, M.A. Saunders, G.W. Stewart, and M.H. Wright. Properties of a representation of a basis for the null space. Mathematical Programming, 33:172-186, 1985.
158. P.E. Gill, W. Murray, M.A. Saunders, and M.H. Wright. User's guide for NPSOL (version 4.0): a Fortran package for nonlinear programming. Technical Report SOL-86-2, Department of Operations Research, Stanford University, Stanford, CA 94305, 1986.
159. P.E. Gill, W. Murray, M.A. Saunders, and M.H. Wright. Constrained nonlinear programming. In G.L. Nemhauser, A.H.G. Rinnooy Kan, and M.J. Todd, editors, Handbooks in Operations Research and Management Science, volume 1: Optimization, chapter 3, pages 171-210. Elsevier Science Publishers B.V., North-Holland, 1989.
160. P.E. Gill, W. Murray, and M.H. Wright. Practical Optimization. Academic Press, New York, 1981.
161. S.T. Glad. Properties of updating methods for the multipliers in augmented Lagrangians. Journal of Optimization Theory and Applications, 28:135-156, 1979.
162. R. Glowinski and Q.-H. Tran. Constrained optimization in reflexion tomography: the augmented Lagrangian method. East-West J. Numer. Math., 1(3):213-234, 1993.
163. J.-L. Goffin. On convergence rates of subgradient optimization methods. Mathematical Programming, 13:329-347, 1977.
164. J.L. Goffin and K.C. Kiwiel. Convergence of a simple sugradient level method. Math. Program., 85:207-211, 1999.
165. D. Goldfarb and A. Idnani. A numerically stable dual method for solving strictly convex quadratic programs. Mathematical Programming, 27:1-33, 1983.
166. D. Goldfarb and M.J. Todd. Linear programming. In G.L. Nemhauser et al., editor, Handbook on Operations Research and Management Science, volume 1, Optimization, pages 73-170. North-Holland, 1989.
167. A.J. Goldman and A.W. Tucker. Polyhedral convex cones. In H.W. Kuhn and A.W. Tucker, editors, Linear inequalities and related systems, pages 1940, Princeton, 1956. Princeton University Press.
168. A.A. Goldstein and J.F. Price. An effective algorithm for minimization. $N u$ merische Mathematik, 10:184-189, 1967.
169. E.G. Gol'shteĭn and N.V. Tretyakov. Modified Lagrangians and Monotone Maps in Optimization. Discrete Mathematics and Optimization. John Wiley \& Sons, New York, 1996.
170. G.H. Golub and C.F. Van Loan. Matrix Computations (second edition). The Johns Hopkins University Press, Baltimore, Maryland, 1989.
171. C.C. Gonzaga. Polynomial affine algorithms for linear programming. Math. Programming, 49:7-21, 1990.
172. C.C. Gonzaga. Path following methods for linear programming. SIAM Review, 34:167-227, 1992.
173. C.C. Gonzaga. A simple presentation of Karmarkar's algorithm. In Workshop on interior point methods, Budapest, 1993.
174. C.C. Gonzaga. The largest step path following algorithm for monotone linear complementarity problems. Mathematical Programming, 76:309-332, 1997.
175. C.C. Gonzaga and J.F. Bonnans. Fast convergence of the simplified largest step path following algorithm. Mathematical Programming series B, 76:95115, 1997.
176. C.C. Gonzaga and R.A. Tapia. On the convergence of the Mizuno-Todd-Ye algorithm to the analytic center of the solution set. SIAM J. Optimization, 7:47-65, 1997.
177. J. Goodman. Newton's method for constrained optimization. Mathematical Programming, 33:162-171, 1985.
178. N. Gould, D. Orban, and Ph.L. Toint. Numerical methods for large-scale nonlinear optimization. In Acta Numerica 2005, pages 299-361. Cambridge University Press, 2005.
179. N.I.M. Gould. On practical conditions for the existence and uniqueness of solutions to the general equality quadratic programming problem. Mathematical Programming, 32:90-99, 1985.
180. G. Gramlich, R. Hettich, and E.W. Sachs. Local convergence of SQP methods in semi-infinite programming. SIAM Journal on Optimization, 5:641-658, 1995.
181. A. Griewank. Evaluating Derivatives - Principles and Techniques of Algorithmic Differentiation. SIAM Publication, 2000.
182. C.B. Gurwitz. Local convergence of a two-piece update of a projected Hessian matrix. SIAM Journal on Optimization, 4:461-485, 1994.
183. W.W. Hager. Stabilized sequential quadratic programming. Computational Optimization and Applications, 12:253-273, 1999.
184. S.-P. Han. Superlinearly convergent variable metric algorithms for general nonlinear programming problems. Mathematical Programming, 11:263-282, 1976.
185. S.-P. Han. A globally convergent method for nonlinear programming. Journal of Optimization Theory and Applications, 22:297-309, 1977.
186. S.-P. Han and O.L. Mangasarian. Exact penalty functions in nonlinear programming. Mathematical Programming, 17:251-269, 1979.
187. P.C. Hansen. Rank-Deficient and Discrete Ill-Posed Problems: Numerical Aspects of Linear Inversion. SIAM, Philadelphia, 1998.
188. M. Held and R. Karp. The traveling salesman problem and minimum spanning trees: Part II. Mathematical Programming, 1(1):6-25, 1971.
189. J. Herskovits. A view on nonlinear optimization. In J. Herskovits, editor, Advances in Structural Optimization, pages 71-116. Kluwer Academic Publishers, 1995.
190. J. Herskovits. Feasible direction interior-point technique for nonlinear optimization. Journal of Optimization Theory and Applications, 99:121-146, 1998.
191. M.R. Hestenes. Multiplier and gradient methods. Journal of Optimization Theory and Applications, 4:303-320, 1969.
192. M.R. Hestenes. Conjugate Direction Methods in Optimization. Number 12 in Applications of Mathematics. Springer-Verlag, 1980.
193. M.R. Hestenes and E. Stiefel. Methods of conjugate gradients for solving linear systems. Journal of Research of the National Bureau of Standards, 49:409-436, 1952.
194. N.J. Higham. Accuracy and Stability of Numerical Algorithms (second edition). SIAM Publication, Philadelphia, 2002.
195. J.-B. Hiriart-Urruty and C. Lemaréchal. Convex Analysis and Minimization Algorithms. Number 305-306 in Grundlehren der mathematischen Wissenschaften. Springer-Verlag, Berlin, 1993.
196. J.-B. Hiriart-Urruty and C. Lemaréchal. Fundamentals of Convex Analysis. Springer-Verlag, Berlin, 2001. Abridged version of Convex analysis and minimization algorithms. I and II [Springer, Berlin, 1993].
197. W. Hoyer. Variants of the reduced Newton method for nonlinear equality constrained optimization problems. Optimization, 17:757-774, 1986.
198. A.D. Ioffe. Necessary and sufficient conditions for a local minimum. 1: a reduction theorem and first order conditions. SIAM Journal on Control and Optimization, 17:245-250, 1979.
199. B. Jansen, C. Roos, and T. Terlaky. The theory of linear programming: skew symmetric self-dual problems and the central path. Optimization, 29:225-233, 1994.
200. A.H.G. Rinnooy Kan and G.T. Timmer. Global optimization. In G.L. Nemhauser, A.H.G. Rinnooy Kan, and M.J. Todd, editors, Handbooks in Operations Research and Management Science, volume 1: Optimization, chapter 9, pages 631-662. Elsevier Science Publishers B.V., North-Holland, 1989.
201. S. Kaniel and A. Dax. A modified Newton's method for unconstrained minimization. SIAM Journal on Numerical Analysis, 16:324-331, 1979.
202. L.V. Kantorovich and G.P. Akilov. Functional Analysis (second edition). Pergamon Press, London, 1982.
203. E. Karas, A. Ribeiro, C. Sagastizábal, and M.V. Solodov. A bundle-filter method for nonsmooth convex constrained optimization. Math. Program. Ser. B, 2006. Accepted for publication.
204. N. Karmarkar. A new polynomial time algorithm for linear programming. Combinatorica, 4:373-395, 1984.
205. J. E. Kelley. The cutting plane method for solving convex programs. J. Soc. Indust. Appl. Math., 8:703-712, 1960.
206. L. Khachian. A polynomial algorithm in linear programming. Soviet Mathematics Doklady, 20:191-194, 1979.
207. L.G. Khachiyan. A polynomial algorithm in linear programming. Doklady Adad. Nauk SSSR, 244:1093-1096, 1979. Trad. anglaise : Soviet Math. Doklady 20(1979), 191-194.
208. K.V. Kim, Yu.E. Nesterov, and B.V. Cherkasskii. An estimate of the effort in computing the gradient. Soviet Math. Dokl., 29:384-387, 1984.
209. K.C. Kiwiel. An exact penalty function algorithm for nonsmooth convex constrained minimization problems. IMA J. Numer. Anal., 5(1):111-119, 1985.
210. K.C. Kiwiel. Methods of Descent for Nondifferentiable Optimization. Lecture Notes in Mathematics 1133. Springer Verlag, Berlin, 1985.
211. K.C. Kiwiel. A constraint linearization method for nondifferentiable convex minimization. Numerische Mathematik, 51:395-414, 1987.
212. K.C. Kiwiel. A subgradient selection method for minimizing convex functions subject to linear constraints. Computing, 39(4):293-305, 1987.
213. K.C. Kiwiel. Proximity control in bundle methods for convex nondifferentiable minimization. Mathematical Programming, 46(1):105-122, 1990.
214. K.C. Kiwiel. Exact penalty functions in proximal bundle methods for constrained convex nondifferentiable minimization. Math. Programming, 52(2, Ser. B):285-302, 1991.
215. K.C. Kiwiel, T. Larsson, and P.O. Lindberg. The efficiency of ballstep subgradient level methods for convex optimization. Mathematics of Operations Research, 24(1):237-254, 1999.
216. V. Klee and G.L. Minty. How good is the simplex algorithm ? In O. Shisha, editor, Inequalities III, pages 159-175. Academic Press, New York, 1972.
217. M. Kojima, N. Megiddo, T. Noma, and A. Yoshise. A unified approach to interior point algorithms for linear complementarity problems. Number 538 in Lecture Notes in Computer Science. Springer Verlag, Berlin, 1991.
218. K. Kortanek and Z. Jishan. New purification algorithms for linear programming. Naval Research Logistics, 35:571-583, 1988.
219. F.-S. Kupfer. An infinite-dimensional convergence theory for reduced SQP methods in Hilbert space. SIAM Journal on Optimization, 6:126-163, 1996.
220. J. Kyparisis. On uniqueness of Kuhn-Tucker multipliers in nonlinear programming. Mathematical Programming, 32:242-246, 1985.
221. L. Lasdon. Optimization Theory for Large Systems. Macmillan Series in Operations Research, 1970.
222. C.T. Lawrence and A.L. Tits. Nonlinear equality constraints in feasible sequential quadratic programming. Optimization Methods and Software, 6:265282, 1996.
223. C.T. Lawrence and A.L. Tits. Feasible sequential quadratic programming for finely discretized problems from SIP. In R. Reemtsen and J.-J. Rückmann, editors, Semi-infinite Programming, pages 159-193. Kluwer Academic Publishers B.V., 1998.
224. C.T. Lawrence and A.L. Tits. A computationally efficient feasible sequential quadratic programming algorithm. SIAM Journal on Optimization, 11:10921118, 2001.
225. F. Leibfritz and E.W. Sachs. Inexact SQP interior point methods and large scale optimal control problems. SIAM Journal on Optimization, 38:272-293, 1999.
226. C. Lemaréchal. An algorithm for minimizing convex functions. In J.L. Rosenfeld, editor, Information Processing '74, pages 552-556. North Holland, 1974.
227. C. Lemaréchal. An extension of Davidon methods to nondifferentiable problems. Mathematical Programming Study, 3:95-109, 1975.
228. C. Lemaréchal. A view of line-searches. In A. Auslender, W. Oettli, and J. Stoer, editors, Optimization and Optimal Control, number 30 in Lecture Notes in Control and Information Science, pages 59-78. Springer-Verlag, Heidelberg, 1981.
229. C. Lemaréchal and R. Mifflin. Global and superlinear convergence of an algorithm for one-dimensional minimzation of convex functions. Mathematical Programming, 24:241-256, 1982.
230. C. Lemaréchal, A.S. Nemirovskii, and Yu.E. Nesterov. New variants of bundle methods. Mathematical Programming, 69:111-148, 1995.
231. C. Lemaréchal, F. Oustry, and C. Sagastizábal. The $\mathcal{U}$-Lagrangian of a convex function. Transactions of the AMS, 352(2):711-729, 2000.
232. C. Lemaréchal, F. Pellegrino, A. Renaud, and C. Sagastizábal. Bundle methods applied to the unit-commitment problem. In J. Doležal and J. Fidler, editors, System Modelling and Optimization, pages 395-402. Chapman and Hall, 1996.
233. C. Lemaréchal and C. Sagastizábal. An approach to variable metric bundle methods. In J. Henry and J-P. Yvon, editors, Systems Modelling and Optimization, number 197 in Lecture Notes in Control and Information Sciences, pages 144-162. Springer Verlag, 1994.
234. C. Lemaréchal and C. Sagastizábal. More than first-order developments of convex functions: primal-dual relations. Journal of Convex Analysis, 3(2):114, 1996.
235. C. Lemaréchal and C. Sagastizábal. Practical aspects of the Moreau-Yosida regularization: theoretical preliminaries. SIAM Journal on Optimization, 7(2):367-385, 1997.
236. C. Lemaréchal and C. Sagastizábal. Variable metric bundle methods: from conceptual to implementable forms. Mathematical Programming, 76(3):393410, 1997.
237. K. Levenberg. A method for the solution of certain nonlinear problems in least squares. Quart. Appl. Math., 2:164-168, 1944.
238. D.C. Liu and J. Nocedal. On the limited memory BFGS method for large scale optimization. Mathematical Programming, 45:503-520, 1989.
239. D.G. Luenberger. Introduction to Linear and Nonlinear Programming (second edition). Addison-Wesley, Reading, USA, 1984.
240. L. Lukšan and J. Vlček. A bundle-Newton method for nonsmooth unconstrained minimization. Math. Programming, 83(3, Ser. A):373-391, 1998.
241. L. Lukšan and J. Vlček. Globally convergent variable metric method for convex nonsmooth unconstrained minimization. J. Optim. Theory Appl., 102(3):593-613, 1999.
242. Z.-Q. Luo, J.-S. Pang, and D. Ralph. Mathematical Programs with Equilibrium Constraints. Cambridge University Press, 1996.
243. Y. Maday, D. Meiron, A.T. Patera, and E.M. Ronquist. Analysis of iterative methods for the steady and unsteady Stokes problem: application to spectral element discretizations. SIAM Journal on Scientific Computing, 14:310-337, 1993.
244. J.H. Maddocks. Restricted quadratic forms, inertia theorems, and the Schur complement. Linear Algebra and its Applications, 108:1-36, 1988.
245. Ph. Mahey, S. Oualibouch, and Pham Dinh Tao. Proximal decomposition on the graph of a maximal monotone operator. SIAM J. Optim., 5(2):454-466, 1995.
246. O.L. Mangasarian and S. Fromovitz. The Fritz John necessary optimality conditions in the presence of equality and inequality constraints. Journal of Mathematical Analysis and Applications, 17:37-47, 1967.
247. N. Maratos. Exact penalty function algorithms for finite dimensional and control optimization problems. PhD thesis, Imperial College, London, 1978.
248. D.W. Marquardt. An algorithm for least-squares estimation of nonlinear parameters. J. Soc. Indust. Appl. Math., 11:431-441, 1963.
249. Mathworks. The Matlab distributed computing engine. http://www. mathworks.com/.
250. D.Q. Mayne and E. Polak. A superlinearly convergent algorithm for constrained optimization problems. Mathematical Programming Study, 16:45-61, 1982.
251. K.A. McShane. A superlinearly convergent $O(\sqrt{n} L)$ iteration interior-point algorithms for linear programming and the monotone linear complementarity problem. SIAM J. Optimization, 4:247-261, 1994.
252. R. Mifflin. An algorithm for constrained optimization with semi-smooth functions. Mathematics of Operations Research, 2(2):191-207, 1977.
253. R. Mifflin. A modification and an extension of Lemaréchal's algorithm for nonsmooth minimization. Mathematical Programming Study, 17:77-90, 1982.
254. R. Mifflin. A quasi-second-order proximal bundle algorithm. Mathematical Programming, 73(1):51-72, 1996.
255. R. Mifflin and C. Sagastizábal. Proximal points are on the fast track. Journal of Convex Analysis, 9(2):563-579, 2002.
256. R. Mifflin and C. Sagastizábal. Primal-Dual Gradient Structured Functions: second-order results; links to epi-derivatives and partly smooth functions. SIAM Journal on Optimization, 13(4):1174-1194, 2003.
257. R. Mifflin and C. Sagastizábal. A $\mathcal{V} U$-proximal point algorithm for minimization. Math. Program., 104(2-3):583-608, 2005.
258. R. Mifflin, D.F. Sun, and L.Q. Qi. Quasi-Newton bundle-type methods for nondifferentiable convex optimization. SIAM Journal on Optimization, 8(2):583-603, 1998.
259. S. Mizuno. A new polynomial time method for a linear complementarity problem. Mathematical Programming, 56:31-43, 1992.
260. S. Mizuno. A superlinearly convergent infeasible-interior-point algorithm for geometrical LCPs without a strictly complementary condition. Mathematics of Operations Research, 21:382-400, 1996.
261. S. Mizuno, F. Jarre, and J. Stoer. A unified approach to infeasible-interiorpoint algorithms via geometrical linear complementarity problems. J. Applied Mathematics and Optimization, 33:315-341, 1994.
262. S. Mizuno, M.J. Todd, and Y. Ye. On adaptative step primal-dual interiorpoint algorithms for linear programming. Mathematics of Operations Research, 18:964-981, 1993.
263. R.D.C. Monteiro and T. Tsuchiya. Limiting behavior of the derivatives of certain trajectories associated with a monotone horizontal linear complementarity problem. Mathematics of Operations Research, 21:793-814, 1996.
264. J.J. Moré. Recent developments in algorithms and software for trust region methods. In A. Bachem, M. Grötschel, and B. Korte, editors, Mathematical Programming, the State of the Art, pages 258-287. Springer-Verlag, Berlin, 1983.
265. J.J. Moré and D.J. Thuente. Line search algorithms with guaranteed sufficient decrease. ACM Transactions on Mathematical Software, 20:286-307, 1994.
266. J.J. Moré and G. Toraldo. Algorithms for bound constrained quadratic programming problems. Numerische Mathematik, 55:377-400, 1989.
267. J.J. Moré and S.J. Wright, editors. Optimization Software Guide, volume 14 of Frontiers in Applied Mathematics. SIAM Publications, 1993.
268. J.J. Moreau. Proximité et dualité dans un espace hilbertien. Bulletin de la Société Mathématique de France, 93:273-299, 1965.
269. W. Murray and M. L. Overton. Steplength algorithms for minimizing a class of nondifferentiable functions. Computing, 23(4):309-331, 1979.
270. W. Murray and F.J. Prieto. A sequential quadratic programming algorithm using an incomplete solution of the subproblem. SIAM Journal on Optimization, 5:590-640, 1995.
271. W. Murray and M.H. Wright. Projected Lagrangian methods based on the trajectories of penalty and barrier functions. Technical Report SOL-78-23, Department of Operations Research, Stanford University, Stanford, CA 94305, 1978.
272. A.S. Nemirovskii and D. Yudin. Problem Complexity and Method Efficiency in Optimization. Wiley-Interscience Series in Discrete Mathematics, 1983. (Original Russian: Nauka, 1979).
273. Y.E. Nesterov and A.S. Nemirovskii. Interior-Point Polynomial Algorithms in Convex Programming. Number 13 in SIAM Studies in Applied Mathematics. SIAM, Philadelphia, 1994.
274. J. Von Neumann and O. Morgenstein. Theory of games and economic behavior. Princeton University Press, Princeton, 1944.
275. J. Nocedal. Updating quasi-Newton matrices with limited storage. Mathematics of Computation, 35:773-782, 1980.
276. J. Nocedal and M.L. Overton. Projected Hessian updating algorithms for nonlinearly constrained optimization. SIAM Journal on Numerical Analysis, 22:821-850, 1985.
277. J. Nocedal and S.J. Wright. Numerical Optimization. Springer Series in Operations Research. Springer, New York, 1999.
278. J.M. Ortega and W.C. Rheinboldt. Iterative Solution of Nonlinear Equations in Several Variables. Academic Press, New York, 1970.
279. A. Ouorou. Epsilon-proximal decomposition method. Math. Program., 99(1, Ser. A):89-108, 2004.
280. U.M. Garcia Palomares and O.L. Mangasarian. Superlinear convergent quasiNewton algorithms for nonlinearly constrained optimization problems. Mathematical Programming, 11:1-13, 1976.
281. E.R. Panier and A.L. Tits. Avoiding the Maratos effect by means of a nonmonotone line search I: General constrained problems. SIAM Journal on Numerical Analysis, 28:1183-1195, 1991.
282. E.R. Panier and A.L. Tits. On combining feasibility, descent and superlinear convergence in inequality constrained optimization. Mathematical Programming, 59(2):261-276, 1993.
283. E.R. Panier, A.L. Tits, and J. Herskovits. A QP-free, globally convergent, locally superlinearly convergent algorithm for inequality constrained optimization. SIAM Journal on Control and Optimization, 26:788-811, 1988.
284. T. Pietrzykowski. An exact potential method for constrained maxima. SIAM Journal on Numerical Analysis, 6:299-304, 1969.
285. E. Polak. Optimization - Algorithms and Consistent Approximations. Number 124 in Applied Mathematical Sciences. Springer, 1997.
286. B.T. Polyak. Introduction to Optimization. Optimization Software, New York, 1987.
287. F.A. Potra. An $O(n L)$ infeasible-interior-point algorithm for LCP with quadratic convergence. Annals of Operations Research, 62:81-102, 1996.
288. M.J.D. Powell. A method for nonlinear constraints in minimization problems. In R. Fletcher, editor, Optimization, pages 283-298. Academic Press, London, New York, 1969.
289. M.J.D. Powell. On the convergence of the variable metric algorithm. Journal of the Institute of Mathematics and its Applications, 7:21-36, 1971.
290. M.J.D. Powell. Some global convergence properties of a variable metric algorithm for minimization without exact line searches. In R.W. Cottle and C.E. Lemke, editors, Nonlinear Programming, number 9 in SIAM-AMS Proceedings. American Mathematical Society, Providence, RI, 1976.
291. M.J.D. Powell. Algorithms for nonlinear constraints that use Lagrangian functions. Mathematical Programming, 14:224-248, 1978.
292. M.J.D. Powell. The convergence of variable metric methods for nonlinearly constrained optimization calculations. In O.L. Mangasarian, R.R. Meyer, and S.M. Robinson, editors, Nonlinear Programming 3, pages 27-63, 1978.
293. M.J.D. Powell. A fast algorithm for nonlinearly constrained optimization calculations. In G.A. Watson, editor, Numerical Analysis Dundee 1977, number 630 in Lecture Notes in Mathematics, pages 144-157. Springer-Verlag, Berlin, 1978.
294. M.J.D. Powell. Nonconvex minimization calculations and the conjugate gradient method. In Lecture Notes in Mathematics 1066, pages 122-141. SpringerVerlag, Berlin, 1984.
295. M.J.D. Powell. Convergence properties of algorithms for nonlinear optimization. SIAM Review, 28:487-500, 1985.
296. M.J.D. Powell. On the quadratic programming algorithm of Goldfarb and Idnani. Mathematical Programming Study, 25:46-61, 1985.
297. M.J.D. Powell. The performance of two subroutines for constrained optimization on some difficult test problems. In P.T. Boggs, R.H. Byrd, and R.B. Schnabel, editors, Numerical Optimization 1984, pages 160-177. SIAM Publication, Philadelphia, 1985.
298. M.J.D. Powell. A view of nonlinear optimization. In J.K. Lenstra, A.H.G. Rinnooy Kan, and A. Schrijver, editors, History of Mathematical Programming, A Collection of Personal Reminiscences, pages 119-125. CWI North-Holland, Amsterdam, 1991.
299. M.J.D. Powell and Y. Yuan. A recursive quadratic programming algorithm that uses differentiable exact penalty functions. Mathematical Programming, 35:265-278, 1986.
300. B.N. Pshenichnyj. Algorithm for a general mathematical programming problem. Kibernetika, 5:120-125, 1970.
301. B.N. Pshenichnyj. The Linearization Method for Constrained Optimization. Number 22 in Computational Mathematics. Springer-Verlag, 1994.
302. B.N. Pshenichnyj and Yu.M. Danilin. Numerical Methods for Extremal Problems. MIR, Moscow, 1978.
303. W. Queck. The convergence factor of preconditioned algorithms of the ArrowHurwicz type. SIAM Journal on Numerical Analysis, 26:1016-1030, 1989.
304. J.K. Reid. A sparsity-exploiting variant of the Bartels-Golub decomposition for linear programming bases. Mathematical Programming, 24:55-69, 1982.
305. P.A. Rey and C.A. Sagastizábal. Dynamical adjustment of the prox-parameter in variable metric bundle methods. Optimization, 51(2):423-447, 2002.
306. S.M. Robinson. A quadratically convergent algorithm for general nonlinear programming problems. Mathematical Programming, 3:145-156, 1972.
307. S.M. Robinson. Perturbed Kuhn-Tucker points and rates of convergence for a class of nonlinear-programming algorithms. Mathematical Programming, 7:1-16, 1974.
308. S.M. Robinson. Generalized equations and their solutions, part II: applications to nonlinear programming. Mathematical Programming Study, 19:200-221, 1982.
309. R.T. Rockafellar. Convex Analysis. Number 28 in Princeton Mathematics Ser. Princeton University Press, Princeton, New Jersey, 1970.
310. R.T. Rockafellar. New applications of duality in convex programming. In Proceedings of the 4th Conference of Probability, Brasov, Romania, pages 7381, 1971.
311. R.T. Rockafellar. Augmented Lagrange multiplier functions and duality in nonconvex programming. SIAM Journal on Control, 12:268-285, 1974.
312. R.T. Rockafellar. Augmented Lagrangians and applications of the proximal point algorithm in convex programming. Mathematics of Operations Research, 1:97-116, 1976.
313. R.T. Rockafellar. Monotone operators and the proximal point algorithm. SIAM Journal on Control and Optimization, 14:877-898, 1976.
314. R.T. Rockafellar and R.J.-B. Wets. Variational Analysis. Springer Verlag, Heidelberg, 1998.
315. T. Rusten and R. Winthier. A preconditioned iterative method for saddle point problems. SIAM Journal on Matrix Analysis and Applications, 13:887904, 1992.
316. A. Ruszczyński. Decomposition methods in stochastic programming. In Mathematical Programming, volume 79, 1997.
317. C. Sagastizábal and M.V. Solodov. On the relation between bundle methods for maximal monotone inclusions and hybrid proximal point algorithms. In Inherently parallel algorithms in feasibility and optimization and their applications (Haifa, 2000), volume 8 of Stud. Comput. Math., pages 441-455. North-Holland, Amsterdam, 2001.
318. C. Sagastizábal and M.V. Solodov. An infeasible bundle method for nonsmooth convex constrained optimization without a penalty function or a filter. SIAM Journal on Optimization, 16(1):146-169, 2005.
319. R. Saigal. Linear Programming: A Modern Integrated Analysis. Kluwer Academic Publishers, Boston, 1995.
320. R.W.H. Sargent. The development of SQP algorithm for nonlinear programming. In L.T. Biegler, T.F. Coleman, A.R. Conn, and F.N. Santosa, editors, Large-Scale Optimization with Applications, part II: Optimal design and Control, pages 1-19. IMA Vol. Math. Appl. 93, 1997.
321. R.W.H. Sargent and M. Ding. A new SQP algorithm for large-scale nonlinear programming. SIAM Journal on Optimization, 11:716-747, 2000.
322. K. Schittkowski. The nonlinear programming method of Wilson, Han and Powell with an augmented Lagrangian type line search function, Part 1: convergence analysis. Numerische Mathematik, 38:83-114, 1981.
323. K. Schittkowski. NLPQL: a FORTRAN subroutine solving constrained nonlinear programming problems. Annals of Operations Research, 5:485-500, 1985.
324. K. Schittkowski. Solving nonlinear programming problems with very many constraints. Optimization, 25:179-196, 1992.
325. K. Schittkowski. Numerical Data Fitting in Dynamical Systems. Kluwer Academic Press, Dordrecht, 2002.
326. H. Schramm and J. Zowe. A version of the bundle idea for minimizing a nonsmooth function: conceptual idea, convergence analysis, numerical results. SIAM Journal on Optimization, 2(1):121-152, 1992.
327. Scilab. A free scientific software package. http://www.scilab.org/.
328. D.F. Shanno. Conjugate gradient methods with inexact searches. Mathematics of Operations Research, 3:244-256, 1978.
329. D.F. Shanno and K.H. Phua. Algorithm 500, minimization of unconstrained multivariate functions. ACM Transactions on Mathematical Software, 2:8794, 1976.
330. A. Shapiro and J. Sun. Some properties of the augmented Lagrangian in cone constrained optimization. Mathematics of Operations Research, 29:479-491, 2004.
331. N. Shor. Utilization of the operation of space dilatation in the minimization of convex functions. Kibernetica, 1:6-12, 1970. (English translation: Cybernetics, 6, 7-15).
332. N.Z. Shor. Minimization methods for non-differentiable functions. Springer Verlag, Berlin, 1985.
333. D. Silvester and A. Wathen. Fast iterative solution of stabilized Stokes systems part II: using general block preconditioners. SIAM Journal on Numerical Analysis, 31:1352-1367, 1994.
334. M. Slater. Lagrange multipliers revisited: a contribution to non-linear programming. Cowles Commission Discussion Paper, Math. 403, 1950.
335. M. V. Solodov. A class of decomposition methods for convex optimization and monotone variational inclusions via the hybrid inexact proximal point framework. Optim. Methods Softw., 19(5):557-575, 2004.
336. M. V. Solodov and B. F. Svaiter. A hybrid approximate extragradientproximal point algorithm using the enlargement of a maximal monotone operator. Set-Valued Anal., 7(4):323-345, 1999.
337. M. V. Solodov and B. F. Svaiter. A hybrid projection-proximal point algorithm. J. Convex Anal., 6(1):59-70, 1999.
338. M. V. Solodov and B. F. Svaiter. A unified framework for some inexact proximal point algorithms. Numerical Functional Analysis and Optimization, 22:1013-1035, 2001.
339. B. Speelpenning. Compiling fast partial derivatives of functions given by algorithms. PhD thesis, Department of Computer Science, University of Illinois at Urbana-Champaign, Urbana, IL 61801, 1980.
340. P. Spellucci. Numerische Verfahren der nichtlinearen Optimierung. Birkhäuser, 1993.
341. P. Spellucci. A new technique for inconsistent problems in the SQP method. Mathematical Methods of Operations Research, 47:355-400, 1998.
342. P. Spellucci. An SQP method for general nonlinear programs using only equality constrained subproblems. Mathematical Programming, 82:413-448, 1998.
343. J. E. Spingarn. Partial inverse of a monotone operator. Appl. Math. Optim., 10(3):247-265, 1983.
344. M. Spivak. A Comprehensive Introduction to Differential Geometry. Publish or Perish, 1979.
345. T. Steihaug. The conjugate gradient method and trust regions in large scale optimization. SIAM Journal on Numerical Analysis, 20:626-637, 1983.
346. R.A. Tapia. Diagonalized multiplier methods and quasi-Newton methods for constrained optimization. Journal of Optimization Theory and Applications, 22:135-194, 1977.
347. R.A. Tapia. On secant updates for use in general constrained optimization. Mathematics of Computation, 51:181-202, 1988.
348. T. Terlaky, editor. Interior Point Methods of Mathematical Programming. Kluwer Academic Publishers, Boston, 1996.
349. T. Terlaky, J. P. Vial, and K. Roos. Theory and algorithms for linear optimization: an interior point approach. Wiley intersciences, New York, 1997.
350. M.J. Todd. On convergence properties of algorithms for unconstrained minimization. IMA Journal of Numerical Analysis, 9(3):435-441, 1989.
351. K. Tone. Revisions of constraint approximations in the successive QP method for nonlinear programming problems. Mathematical Programming, 26:144152, 1983.
352. P. Tseng. Alternating projection-proximal methods for convex programming and variational inequalities. SIAM J. Optim., 7(4):951-965, 1997.
353. R.J. Vanderbei. Linear Programming: Foundations and extensions. Kluwer Academic Publishers, Boston, 1997.
354. S.A. Vavasis. Nonlinear Optimization - Complexity Issues. Oxford University Press, New York, 1991.
355. R. Verfürth. A combined conjugate gradient-multigrid algorithm for the numerical solution of the Stokes problem. IMA Journal of Numerical Analysis, 4:441-455, 1984.
356. K. Veselić. Finite catenary and the method of Lagrange. SIAM Review, 37:224-229, 1995.
357. M. Wagner and M.J. Todd. Least-change quasi-Newton updates for equalityconstrained optimization. Mathematical Programming, 87:317-350, 2000.
358. A. Wathen and D. Silvester. Fast iterative solution of stabilized Stokes systems part I: using simple diagonal preconditioners. SIAM Journal on Numerical Analysis, 30:630-649, 1993.
359. R.B. Wilson. A simplicial algorithm for concave programming. PhD thesis, Graduate School of Business Administration, Harvard University, Cambridge, MA, USA, 1963.
360. P. Wolfe. A duality theorem for nonlinear programming. Quarterly Applied Mathematics, 19:239-244, 1961.
361. P. Wolfe. Convergence conditions for ascent methods. SIAM Review, 11:226235, 1969.
362. P. Wolfe. A method of conjugate subgradients for minimizing nondifferentiable functions. Mathematical Programming Study, 3:145-173, 1975.
363. H. Wolkowicz, R. Saigal, and L. Vandenberghe, editors. Handbook of Semidefinite Programming - Theory, Algorithms, and Applications. Kluwer Academic Publishers, 2000.
364. S.J. Wright. A path-following infeasible-interior-point algorithm for linear complementarity problems. Optimization Methods and Software, 2:79-106, 1993.
365. S.J. Wright. An infeasible interior point algorithm for linear complementarity problems. Mathematical Programming, 67:29-52, 1994.
366. S.J. Wright. Primal-dual interior-point methods. SIAM, Philadelphia, 1996.
367. S.J. Wright. Superlinear convergence of a stabilized SQP method to a degenerate solution. Computational Optimization and Applications, 11:253-275, 1998.
368. Y. Xie and R.H. Byrd. Practical update criteria for reduced Hessian SQP: global analysis. SIAM Journal on Optimization, 9:578-604, 1999.
369. X. Xu, P.F. Hung, and Y. Ye. A simplification of the homogeneous and self-dual linear programming algorithm and its implementation. Annals of Operations Research, 62:151-172, 1996.
370. Y. Ye. Interior point algorithms. Wiley-Interscience Series in Discrete Mathematics and Optimization. John Wiley \& Sons Inc., New York, 1997.
371. Y. Ye. On homogeneous and self-dual algorithm for LCP. Mathematical Programming, 76:211-222, 1997.
372. Y. Ye, M.J. Todd, and S. Mizuno. An $O(\sqrt{n} L)$-iteration homogeneous and self-dual linear programming algorithm. Mathematics of Operations Research, 19:53-67, 1994.
373. Y. Yuan. An only 2 -step $Q$-superlinearly convergence example for some algorithms that use reduced Hessian informations. Mathematical Programming, 32:224-231, 1985.
374. W.I. Zangwill. Non-linear programming via penalty functions. Management Science, 13:344-358, 1967.
375. Y. Zhang. On the convergence of a class of infeasible interior-point methods for the horizontal linear complementarity problem. SIAM J. Optimization, 4:208-227, 1994.
376. G. Zoutendijk. Nonlinear programming, computational methods. In J. Abadie, editor, Integer and Nonlinear Programming, pages 37-86. North-Holland, Amsterdam, 1970.


[^0]:    ${ }^{1}$ See exercise 14.1 for an example, in which $f$ is concave. When $f$ is strongly convex and has a bounded Hessian, one can get convergence with line-search along the direction computed by (14.2). When $f$ is nonconvex, convergence can still be obtained with line-search and the truncated SQP algorithm. This will be clearer with the concepts developed in chapter 17. Nevertheless, as this is shown below, the step computed by (14.2) neglects an important part of the "curvature" of problem $\left(P_{E}\right)$.

