# Algorithmic Results for Clustering and Refined Physarum Analysis 

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# Colloquium Details 

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## Abstract

In the first part of this thesis, we study the Binary $\ell_{0}$-Rank- $k$ problem which given a binary matrix $A$ and a positive integer $k$, seeks to find a rank- $k$ binary matrix $B$ minimizing the number of non-zero entries of $A-B$. A central open question is whether this problem admits a polynomial time approximation scheme. We give an affirmative answer to this question by designing the first randomized almost-linear time approximation scheme for constant $k$ over the reals, $\mathbb{F}_{2}$, and the Boolean semiring. In addition, we give novel algorithms for important variants of $\ell_{0}$-low rank approximation.

The second part of this dissertation, studies a popular and successful heuristic, known as Approximate Spectral Clustering (ASC), for partitioning the nodes of a graph $G$ into clusters with small conductance. We give a comprehensive analysis, showing that ASC runs efficiently and yields a good approximation of an optimal $k$-way node partition of $G$.

In the final part of this thesis, we present two results on slime mold computations: i) the continuous undirected Physarum dynamics converges for undirected linear programs with a non-negative cost vector; and ii) for the discrete directed Physarum dynamics, we give a refined analysis that yields strengthened and close to optimal convergence rate bounds, and shows that the model can be initialized with any strongly dominating point.

## Zusammenfassung

Im ersten Teil dieser Arbeit untersuchen wir das Binary $\ell_{0}$-Rank- $k$ Problem. Hier sind eine binäre Matrix $A$ und eine positive ganze Zahl $k$ gegeben und gesucht wird eine binäre Matrix $B$ mit Rang $k$, welche die Anzahl von nicht null Einträgen in $A-B$ minimiert. Wir stellen das erste randomisierte, nahezu lineare Aproximationsschema vor konstantes $k$ über die reellen Zahlen, $\mathbb{F}_{2}$ und den Booleschen Semiring. Zusätzlich erzielen wir neue Algorithmen für wichtige Varianten der $\ell_{0}$-low rank Approximation.

Der zweite Teil dieser Dissertation beschäftigt sich mit einer beliebten und erfolgreichen Heuristik, die unter dem Namen Approximate Spectral Cluster (ASC) bekannt ist. ASC partitioniert die Knoten eines gegeben Graphen $G$ in Cluster kleiner Conductance. Wir geben eine umfassende Analyse von ASC, die zeigt, dass ASC eine effiziente Laufzeit besitzt und eine gute Approximation einer optimale $k$-Weg-Knoten Partition für $G$ berechnet.

Im letzten Teil dieser Dissertation präsentieren wir zwei Ergebnisse über Berechnungen mit Hilfe von Schleimpilzen: i) die kontinuierliche ungerichtete Physarum Dynamik konvergiert für ungerichtete lineare Programme mit einem nicht negativen Kostenvektor; und ii) für die diskrete gerichtete Physikum Dynamik geben wir eine verfeinerte Analyse, die stärkere und beinahe optimale Schranken für ihre Konvergenzraten liefert und zeigt, dass das Model mit einem beliebigen stark dominierender Punkt initialisiert werden kann.

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## Preface

In the first part of this thesis, we study NP-Hard variants of low rank approximation that are natural for problems with no underlying metric. We consider the Binary $\ell_{0}$-Rank- $k$ problem which given a binary matrix $A \in\{0,1\}^{m \times n}$ with $m \geqslant n$ and a positive integer $k$, seeks to find a rank- $k$ binary matrix $B \in\{0,1\}^{m \times n}$ minimizing $\|A-B\|_{0}$, where $\|\cdot\|_{0}$ denotes the number of non-zero entries. This problem is known under many different names in different areas of computer science: $\ell_{0}$-low rank approximation in computational linear algebra, constrained binary matrix factorization in data mining, Boolean factor analysis in machine learning, and matrix rigidity in computational complexity theory. A central open question is whether the Binary $\ell_{0}$-Rank- $k$ problem admits a polynomial time approximation scheme (PTAS).

In a joint work $\left[\mathrm{BBB}^{+} 18, \mathrm{BBB}^{+} 19\right]$ with Frank Ban, Vijay Bhattiprolu, Karl Bringmann, Euiwoong Lee and David P. Woodruff, we give an affirmative answer to this question by designing the first PTAS for the more general problem of $\ell_{p}$-Rank- $k$ for any $p \in[0,2)$. Approximately, Frank Ban and David P. Woodruff contributed the PTAS for $p \in(0,2)$, Vijay Bhattiprolu and Euiwoong Lee contributed the hardness results for $p \in(1,2)$ and finite fields of constant size, and Karl Bringmann and I contributed the following PTAS for what we call the Generalized Binary $\ell_{0}$-Rank- $k$ problem.

In particular, we give the first randomized almost-linear time approximation scheme for the Generalized Binary $\ell_{0}$-Rank- $k$ problem. Our algorithm finds a $(1+\varepsilon)$-approximation in time $(2 / \varepsilon)^{2^{O(k)} / \varepsilon^{2}} \cdot m n^{1+o(1)}$, where $o(1)$ hides a factor $(\log \log n)^{1.1} / \log n$. This yields the first PTAS for the Binary $\ell_{0}$-Rank- $k$ problem for constant $k$ over the reals, $\mathbb{F}_{2}$, and the Boolean semiring. Even for the special case of rank $k=1$ no PTAS was known before. Our algorithmic techniques crucially rely on the fact that the Generalized Binary $\ell_{0}$-Rank- $k$ problem is equivalent to a variant of clustering problem with constrained centers, and the crux of our work is to extend existing algorithmic techniques for unconstrained clustering to the more general setting of clustering with constrained centers.

Further, in a joint work [BKW17a, BKW17b] with Karl Bringmann and David P. Woodruff, we give novel algorithms for the following important variants of $\ell_{0}$-low rank approximation:
(a) a polynomial time algorithm for the Reals $\ell_{0}$-Rank- $k$ problem, if we allow for a bicriteria solution, that outputs a matrix with larger rank $O(k \log (n / k))$ and approximation factor $O\left(k^{2} \log (n / k)\right)$;
(b) a sublinear time $(2+\varepsilon)$-approximation algorithm for the Reals $\ell_{0}$-Rank- 1 problem;
(c) a sublinear time $(1+O(\psi))$-approximation algorithm for the Binary $\ell_{0}$-Rank-1 problem, where $\psi=\|A\|_{0} / \mathrm{OPT}$, and a matching sample complexity lower bound;
(d) an exact algorithm running in time $2^{O\left(\mathrm{OPT} / \sqrt{\left.\|A\|_{0}\right)}\right.} \operatorname{poly}(m n)$ for the Binary $\ell_{0}$-Rank-1 problem.

The second part of this dissertation, studies a popular and successful heuristic, known as Approximate Spectral Clustering (ASC), for partitioning the nodes of a graph $G$ into clusters for which the ratio of outside connections compared to the volume (sum of degrees) is small. ASC consists of the following two subroutines: i) compute an approximate Spectral Embedding via the Power method; and ii) partition the resulting vector set with an approximate $k$-means clustering algorithm. The resulting $k$-means partition naturally induces a $k$-way node partition of $G$.

In a joint work [KM16, KM18] with Kurt Mehlhorn, we give a comprehensive analysis of ASC building on the work of Peng et al. (2017), Boutsidis et al. (2015) and Ostrovsky et al. (2013). We show that ASC runs efficiently and yields a good approximation of an optimal $k$-way node partition of $G$. Moreover, we strengthen the quality guarantees of a structural result of Peng et al. by a factor of $k$, and simultaneously weaken the eigenvalue gap assumption. Further, we demonstrate that ASC finds a $k$-way node partition of $G$ with the strengthened quality guarantees.

The final part of this thesis is a joint work $\left[\mathrm{BBK}^{+} 17, \mathrm{BBK}^{+} 18\right]$ with Ruben Becker, Vincenzo Bonifaci, Andreas Karrenbauer and Kurt Mehlhorn. We present two results on slime mold computations:
(1) Bonifaci, Mehlhorn and Varma (2012) showed that the continuous undirected Physarum dynamics, a system of differential equations, converges for the shortest path problems. We demonstrate that the dynamics actually converges for a much wider class of problem, namely undirected linear programs with a non-negative cost vector.
(2) Combinatorial optimization researchers took the dynamics describing slime mold behavior as an inspiration for an optimization method, and in a research line culminating with the work of Straszak and Vishnoi (2016) showed that its discretization initialized with a feasible point can approximately solve linear programs with positive cost vector. We give a refined analysis that yields strengthened and close to optimal convergence rate bounds and shows that the discrete directed Physarum dynamics can be initialized with any strongly dominating point.

During my doctoral studies, I was also involved in a project on sparsifying Random Walk Laplacian (RWL) matrices. [JK15] gives density-independent algorithms for special families of RWL matrices, and this is a joint work with Gorav Jindal. [JKPS17a, JKPS17b] establishes a density-independent algorithm for general RWL matrices, and this is a joint work with Gorav Jindal, Richard Peng and Saurabh Sawlani.

The following list consists of all publications during my doctoral studies:

## Journal Versions:

[ $\mathrm{BBK}^{+}{ }^{18]}$ Ruben Becker, Vincenzo Bonifaci, Andreas Karrenbauer, Pavel Kolev, and Kurt Mehlhorn. Two results on slime mold computations. Theoretical Computer Science, 2018
[KM18] Pavel Kolev and Kurt Mehlhorn. Approximate spectral clustering: Efficiency and guarantees. CoRR, abs/1509.09188, 2018. Submitted. A preliminary version of this paper was presented at the 24th Annual European Symposium on Algorithms (ESA 2016)

## Conference Versions:

$\left[\right.$ BBB $^{+}{ }^{19]}$ Frank Ban, Vijay Bhattiprolu, Karl Bringmann, Pavel Kolev, Euiwoong Lee, and David P. Woodruff. A PTAS for $\mathrm{l}_{\mathrm{p}}$-low rank approximation. In Proceedings of the 30th Annual ACM-SIAM Symposium on Discrete Algorithms, SODA 2019, Westin San Diego, San Diego, California, USA, January 6-9, 2019. To appear
[BKW17a] Karl Bringmann, Pavel Kolev, and David P. Woodruff. Approximation algorithms for $\mathrm{l}_{0}-$ low rank approximation. In Advances in Neural Information Processing Systems 30: Annual Conference on Neural Information Processing Systems, NIPS 2017, 4-9 December 2017, Long Beach, CA, USA, pages 6651-6662, 2017
[KM16] Pavel Kolev and Kurt Mehlhorn. A note on spectral clustering. In 24th Annual European Symposium on Algorithms, ESA 2016, August 22-24, 2016, Aarhus, Denmark, pages 57:1-57:14, 2016
[JKPS17a] Gorav Jindal, Pavel Kolev, Richard Peng, and Saurabh Sawlani. Density Independent Algorithms for Sparsifying k-Step Random Walks. In Approximation, Randomization, and Combinatorial Optimization. Algorithms and Techniques (APPROX/RANDOM 2017), volume 81 of Leibniz International Proceedings in Informatics (LIPIcs), Dagstuhl, Germany, 2017. Schloss Dagstuhl-Leibniz-Zentrum fuer Informatik

## ArXiv Versions:

$\left[\mathrm{BBB}^{+}\right.$18] Frank Ban, Vijay Bhattiprolu, Karl Bringmann, Pavel Kolev, Euiwoong Lee, and David P. Woodruff. A PTAS for $\mathrm{l}_{\mathrm{p}}$-low rank approximation. CoRR, abs/1807.06101, 2018
[BKW17b] Karl Bringmann, Pavel Kolev, and David P. Woodruff. Approximation algorithms for $\mathrm{l}_{0}$-low rank approximation. CoRR, abs/1710.11253, 2017. Full version of [BKW17a]
$\left[\mathrm{BBK}^{+}{ }^{17}\right]$ Ruben Becker, Vincenzo Bonifaci, Andreas Karrenbauer, Pavel Kolev, and Kurt Mehlhorn. Two results on slime mold computations. CoRR, abs/1707.06631, 2017
[JKPSS17b] Gorav Jindal, Pavel Kolev, Richard Peng, and Saurabh Sawlani. Density independent algorithms for sparsifying k-step random walks. CoRR, abs/1702.06110, 2017
[JK15] Gorav Jindal and Pavel Kolev. An efficient parallel algorithm for spectral sparsification of laplacian and sddm matrix polynomials. CoRR, abs/1507.07497, 2015

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## Part I

## Low Rank Matrix Approximation

## Proceedings

This part is based on an article $\left[\mathrm{BBB}^{+} 19\right]$ that is accepted at $S O D A^{\prime} 19$ and on an article [BKW17a] that was presented at NIPS'17. The full ArXiv versions are available online [BKW17b, $\mathrm{BBB}^{+}$18].

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## Chapter 1

## Introduction

Low rank approximation of an $m \times n$ matrix $A$ is an extremely well-studied problem, where the goal is to replace the matrix $A$ with a rank- $k$ matrix $A^{\prime}$ which well-approximates $A$, in the sense that $\left\|A-A^{\prime}\right\|$ is small under some measure $\|\cdot\|$. Since any rank- $k$ matrix $A^{\prime}$ can be written as $U \cdot V$, where $U$ is $m \times k$ and $V$ is $k \times n$, it suffices to store the $k(m+n)$ entries of $U$ and $V$, which is a significant reduction compared to the $m n$ entries of $A$. Furthermore, computing $A^{\prime} x=U(V x)$ takes time $O(k(m+n))$, which is much less than the time $O(m n)$ for computing $A x$. We refer the reader to several surveys [KV09, Mah11, Woo14] for references to the many results on low rank approximation.
$\ell_{0}$-low rank approximation When the measure $\left\|A-A^{\prime}\right\|_{0}$ is the number of non-zero entries, we seek a rank-k matrix $A^{\prime}$ for which the number of entries $(i, j)$ with $A_{i, j}^{\prime} \neq A_{i, j}$ is as small as possible. This can be seen as the Hamming distance between a matrix $A$ and its best rank $k$ approximation $A^{\prime}$. The $\ell_{0}$-low rank approximation is natural for problems with no underlying metric, since the $\|\cdot\|_{0}$ measure directly answers the following question: if we are allowed to ignore outliers (or anomalies), what is the best low-rank model we can get?

A well-studied case is when $A$ is binary, but $A^{\prime}$ and its factors $U$ and $V$ need not necessarily be binary. This is called unconstrained Binary Matrix Factorization [JPHY14]. There is also a large body of work on the constrained version, in which not only the input matrix $A$ is binary, but we also have the natural restriction that the factors $U, V$ are binary. That is, we study the following problem.

Binary $\ell_{0}$-Rank- $k$ Given a matrix $A \in\{0,1\}^{m \times n}$ and an integer $k$, compute matrices $U \in\{0,1\}^{m \times k}$ and $V \in\{0,1\}^{k \times n}$ minimizing $\|A-U \cdot V\|_{0}$.

Note that we did not yet specify the ground field (or, more generally, semiring) that we are working over, which affects the type of matrix multiplication used in $U \cdot V$. Specifically, for $A^{\prime}=U \cdot V$ we can write the entry $A_{i, j}^{\prime}$ as the inner product of the $i$-th row of $U$ with the $j$-column of $V$ - and the specific inner product function $\langle.,$.$\rangle depends on the ground field.$

Typical choices for the ground field are as follows. Let $x, y \in\{0,1\}^{k}$.
(i) Reals: For the ground field $\mathbb{R}$, we get the standard inner product $\langle x, y\rangle=\sum_{i=1}^{k} x_{i} \cdot y_{i} \in\{0,1, \ldots, k\}$. This is rather unnatural, since no number $\geqslant 2$ can match any entry of $A$.
(ii) Modulo 2: For the ground field $\mathbb{F}_{2}$, the inner product is $\langle x, y\rangle=\bigoplus_{i=1}^{k} x_{i} \cdot y_{i} \in\{0,1\}$. This is used, e.g., in [Yer11, GGYT12, DAJ ${ }^{+}$15, PRF16].
(iii) Boolean: Using the Boolean semiring $\{0,1, \wedge, \vee\}$, the inner product becomes $\langle x, y\rangle=\bigvee_{i=1}^{k} x_{i} \wedge y_{i}=$ $1-\prod_{i=1}^{k}\left(1-x_{i} \cdot y_{i}\right) \in\{0,1\}$. This is used, e.g., in [BV10, DAJ ${ }^{+} 15, \mathrm{MMG}^{+} 08$, SBM03, SH06, VAG07].
Note that for $k=1$ all three inner products are the same.

Clustering with Constrained Centers We now demonstrate that the Binary $\ell_{0}$-Rank- $k$ problem is equivalent to a clustering problem with constrained centers. Consider arbitrary matrices $U \in\{0,1\}^{m \times k}$ and $V \in\{0,1\}^{k \times n}$. Grouping the identical columns of $V$ gives rise to a column partitioning, and let $C_{y}=\left\{j \in[n]: V_{:, j}=y\right\}$ be the set of column indices corresponding to a vector $y$, for every $y \in\{0,1\}^{k}$. Then, the expression $\|A-U \cdot V\|_{0}$ is equivalent to a clustering formulation whose centers are restricted to the set $S_{U} \stackrel{\text { def }}{=}\left\{U \cdot y \in\{0,1\}^{m}: y \in\{0,1\}^{k}\right\}$, and it reads

$$
\|A-U \cdot V\|_{0}=\sum_{j=1}^{n}\left\|A_{:, j}-U \cdot V_{:, j}\right\|_{0}=\sum_{y \in\{0,1\}^{k}} \sum_{j \in C_{y}}\left\|A_{:, j}-U \cdot y\right\|_{0}
$$

Observe that any column of $U \cdot V$ is in $S_{U}$, and thus the choice of columns in $V$ can be seen as selecting constrained centers in $S_{U}$. Formally, we rephrase the Binary $\ell_{0}$-Rank- $k$ problem as follows

$$
\begin{aligned}
\min _{U \in\{0,1\}^{m \times k}, V \in\{0,1\}^{k \times n}}\|A-U \cdot V\|_{0} & =\min _{U \in\{0,1\}^{m \times k}} \sum_{j=1}^{n} \min _{V_{:, j} \in\{0,1\}^{k}}\left\|A_{:, j}-U \cdot V_{:, j}\right\|_{0} \\
& =\min _{U \in\{0,1\}^{m \times k}} \sum_{j=1}^{n} \min _{s \in S_{U}}\left\|A_{:, j}-s\right\|_{0} .
\end{aligned}
$$

This is a clustering problem, where the task is to choose a set of constrained centers $S_{U}$, in order to minimize the total $\ell_{0}$-distance of all columns of $A$ to their closest center in $S_{U}$. The main difference to unconstrained clustering problems is that the constrained centers in $S_{U}$ cannot be chosen independently, since these centers satisfy a certain system of linear equations. Because of this dependence, the techniques used for designing approximation schemes for unconstrained clustering problems are not directly applicable.

Applications The Binary $\ell_{0}$-Rank- $k$ problem arises in many fields of computer science, with different ground fields being important in different areas. In computational linear algebra the problem is known under the umbrella term "low rank matrix approximation" and used for compressing matrices, as described above [KV09, Mah11, Woo14, BKW17a].

In complexity theory the problem is known as "matrix rigidity", introduced by Grigoriev [Gri80] and Valiant [Val77], who showed that constructions of rigid matrices would imply circuit lower bounds, see also [AW17]. The decision problem of matrix rigidity has been studied from the viewpoint of $\operatorname{FPT}\left[\mathrm{FLM}^{+} 17\right]$.

In data mining and machine learning, the problem is known as "(Constrained) Binary Matrix Factorization" or "Boolean Factor Analysis" [Mie, JPHY14, DAJ ${ }^{+}$15, FGP18]. In these areas, the columns of $U$ could correspond to latent topics learned from the database $A$, and, more generally, the problem has numerous applications including latent variable analysis, topic models, association rule mining, clustering, and database tiling [SBM03, SH06, VAG07, MV14, GGYT12]. The usage of the problem to mine discrete patterns has also been applied in bioinformatics to analyze gene expression data [MGT15, SJY09, ZLD ${ }^{+}$10]. In these situations it is natural to consider binary matrices whenever the data is categorical, which is often the case for text data [DAJ ${ }^{+} 15$, MV14, RPG16]. The special and important case in which $A$ is binary and $k=1$ was studied in [KG03, SJY09, JPHY14], as their algorithm for $k=1$ forms the basis for their successful heuristic for general $k$, e.g. the PROXIMUS technique [KG03].

Another special case of Binary $\ell_{0}$-Rank- $k$ is the Biclique Partition problem, see, e.g., [CIK16, FGP18]. Binary $\ell_{0}$-Rank- $k$ over $\mathbb{F}_{2}$ has been applied to Independent Component Analysis of string data in the area of information theory [Yer11, GGYT12, PRF16]. If we drop the requirement that $U$ and $V$ are binary, and use the Frobenius norm as distance measure, we obtain the unconstrained Binary Matrix Factorization [JPHY14], which has applications to association rule mining [KG03], biclustering structure identification [ZLD ${ }^{+} 10$, ZLDZ07], pattern discovery for gene expression [SJY09], digits reconstruction [MGNR06], mining high-dimensional discrete-attribute data [KGR05, KGR06], market based clustering [Li05], and document clustering [ZLDZ07].

An important related problem is robust PCA [CLMW11], in which there is an underlying matrix $A$ that can be written as a low rank matrix $L$ plus a sparse matrix $S$ [CLMW11]. Candès et al. [CLMW11] argue that both the components of $L$ and $S$ are of arbitrary magnitude, and we do not know the locations of the non-zeros in $S$ nor how many there are. Moreover, grossly corrupted observations are common in image processing, web data analysis, and bioinformatics where some measurements are arbitrarily corrupted due to occlusions, malicious tampering, or sensor failures. Specific scenarios include video surveillance, face recognition, latent semantic indexing, and ranking of movies, books, etc. [CLMW11]. These problems have the common theme of being an arbitrary magnitude sparse perturbation to a low rank matrix with no natural underlying metric, and thus the $\ell_{0}$-error measure is appropriate. In order to solve the robust PCA in practice, Candès et al. [CLMW11] relaxed the $\ell_{0}$-error measure to the $\ell_{1}$-norm.

It is of a fundamental importance for theory to understand the algorithmic guarantees and limits for solving the $\ell_{0}$-low rank approximation problem.

Algorithms for Binary $\ell_{0}$-Rank- $k$ Let us first discuss rank $k=1$, in which case the real, $\mathbb{F}_{2}$, and Boolean variants are all equal. A polynomial-time 2-approximation was designed by Shen et al. [SJY09], and simplified by Jiang et al. [JPHY14]. For $k>1$, Dan et al. [DAJ ${ }^{+}$15] presented an $n^{O(k)}$-time $O(k)$ approximation over $\mathbb{F}_{2}$, and an $n^{O(k)}$-time $O\left(2^{k}\right)$-approximation over the Boolean semiring. The work of

Chierichetti et al. [CGK $\left.{ }^{+} 17\right]$ for $\ell_{p}$-low rank approximation, for $p \geqslant 1$, fails to give any approximation factor for $p=0$. Indeed, critical to their analysis is the scale-invariance property of a norm, which does not hold for $p=0$ since $\ell_{0}$ is not a norm. To the best of our knowledge, these are the only known approximation algorithms for this problem.

Hardness of Binary $\ell_{0}$-Rank- $k$ It is well-known that Binary $\ell_{0}$-Rank- $k$ is NP-hard, even for rank $k=1\left[G V 15, \mathrm{DAJ}^{+} 15\right]$. Further, if $k$ is unbounded then even deciding whether matrix $A$ has rank $k$ is NP-hard over the Boolean semiring [Mie]. This suggests that the running time of any approximation algorithm must depend at least exponentially on $k$.

Despite the high interest in the problem in linear algebra, data mining, machine learning, and complexity theory, the approximability of the problem is wide open. In particular, this leaves the following question open. Does the Binary $\ell_{0}-R a n k-k$ problem have a polynomial time approximation scheme (PTAS) for any constant $k$ ?

### 1.1 Generalized Binary $\ell_{0}$-Rank- $k$

The results in this Section are proven in Chapter 2. In order to study the real, $\mathbb{F}_{2}$, and Boolean setting in a unified way, we introduce the following more general problem.

Generalized Binary $\ell_{0}$-Rank- $k$ Given a matrix $A \in\{0,1\}^{m \times n}$ with $m \geqslant n$, an integer $k$, and an inner product function $\langle.,\rangle:.\{0,1\}^{k} \times\{0,1\}^{k} \rightarrow \mathbb{R}$, compute matrices $U \in\{0,1\}^{m \times k}$ and $V \in\{0,1\}^{k \times n}$ minimizing $\|A-U \cdot V\|_{0}$, where the product $U \cdot V$ uses $\langle.,$.$\rangle . An algorithm for the Generalized Binary$ $\ell_{0}$-Rank- $k$ problem is an $\alpha$-approximation, if it outputs matrices $U \in\{0,1\}^{m \times k}$ and $V \in\{0,1\}^{k \times n}$ satisfying $\|A-U \cdot V\|_{0} \leqslant \alpha \cdot \min _{U^{\prime} \in\{0,1\}^{m \times k}, V^{\prime} \in\{0,1\}^{k \times n}}\left\|A-U^{\prime} \cdot V^{\prime}\right\|_{0}$.

As shown above, by choosing an appropriate inner product function $\langle.,$.$\rangle which also runs in time O(k)$, we obtain the Binary $\ell_{0}$-Rank- $k$ problem over the reals, $\mathbb{F}_{2}$, and the Boolean semiring. We assume that the function $\langle.,$.$\rangle can be evaluated in time 2^{O(k)}$, in order to simplify our running time bounds.

Our main result is a randomized almost-linear time approximation scheme for the Generalized Binary $\ell_{0}$-Rank- $k$ problem for any constant $k$. In particular, this yields the first PTAS for the Binary $\ell_{0}$-Rank- $k$ problem for constant $k$ over the reals, $\mathbb{F}_{2}$, and the Boolean semiring. Even for the special case of rank $k=1$ no PTAS was known before.

Theorem 1.1. (PTAS) For any error $\varepsilon \in(0,1 / 2)$, there is a $(1+\varepsilon)$-approximation algorithm for the Generalized Binary $\ell_{0}$-Rank-k problem that runs in time $(2 / \varepsilon)^{2^{O(k)} / \varepsilon^{2}} \cdot m n^{1+o(1)}$ and succeeds with constant probability ${ }^{1}$, where $o(1)$ hides a factor $(\log \log n)^{1.1} / \log n$.

Moreover, we show that our PTAS has a close to optimal running time, in the sense that the runtime of any PTAS for the Generalized Binary $\ell_{0}$-Rank- $k$ problem must depend exponentially on $1 / \varepsilon$ and doubly exponentially on $k$, assuming the Exponential Time Hypothesis (ETH).

Theorem 1.2. (Hardness for Generalized Binary $\ell_{0}$-Rank-k) Assuming the Exponential Time Hypothesis, Generalized Binary $\ell_{0}-R a n k-k$ has no $(1+\varepsilon)$-approximation algorithm in time $2^{1 / \varepsilon^{o(1)}} \cdot 2^{m^{o(1)}}$. Further, for any $\varepsilon \geqslant 0$, Generalized Binary $\ell_{0}$-Rank-k has no $(1+\varepsilon)$-approximation algorithm in time $2^{2^{o(k)}} \cdot 2^{m^{o(1)}}$.

Further, we give a faster algorithm for the Binary $\ell_{0}$-Rank-1 problem with standard inner product. In the following, we assume ${ }^{2}$ that we can access any entry $A_{i, j}$ in constant time, and we can also enumerate all non-zero entries in time $O\left(\|A\|_{0}\right)$.

Theorem 1.3. (PTAS for the Binary $\ell_{0}$-Rank-1 problem with standard inner product) For any $\varepsilon \in(0,1 / 2)$ there is an algorithm that runs in time $(1 / \varepsilon)^{O\left(1 / \varepsilon^{2}\right)} \cdot\left(\|A\|_{0}+m+n\right) \cdot \log ^{3}(m n)$, and outputs vectors $\widetilde{u} \in\{0,1\}^{m}, \widetilde{v} \in\{0,1\}^{n}$ such that w.h.p. ${ }^{3}\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0} \leqslant(1+\varepsilon) \min _{u \in\{0,1\}^{m}, v \in\{0,1\}^{n}}\left\|A-u \cdot v^{T}\right\|_{0}$.

[^0]Note: Our results in Theorem 1.1 and Theorem 1.2 are in submission as of April 2018. Shortly after posting our manuscript $\left[\mathrm{BBB}^{+} 18\right]$ to arXiv on 16 July 2018, we became aware that in an unpublished work Fomin et al. have independently obtained a very similar PTAS for Binary $\ell_{0}$-Rank- $k$ problem. Their manuscript $\left[\mathrm{FGL}^{+} 18\right]$ was posted to arXiv on 18 July 2018. Interestingly, [ $\mathrm{BBB}^{+} 18, \mathrm{FGL}^{+} 18$ ] have independently discovered i) a reduction between the Binary $\ell_{0}$-Rank- $k$ problem and a clustering problem with constrained centers; ii) a structural sampling theorem extending [AS99] which yields a simple but inefficient deterministic PTAS; and iii) an efficient sampling procedure, building on ideas from [KSS04, $\mathrm{ABH}^{+} 05$, ABS10], which gives an efficient randomized PTAS. Notably, by establishing an additional structural result, Fomin et al. $\left[\mathrm{FGL}^{+} 18\right]$ design a faster sampling procedure which yields a randomized PTAS for the Binary $\ell_{0}$-Rank- $k$ problem that runs in linear time $(1 / \varepsilon)^{2^{O(k)} / \varepsilon^{2}} \cdot m n$.

## Further Related Work

Suppose $A \in\{0,1\}^{m \times n}$ and the inner product $\langle.,$.$\rangle maps to \{0,1\}$. Then, matrix $U \cdot V$ has entries in $\{0,1\}$, and hence matrix $B \stackrel{\text { def }}{=} A-U \cdot V$ has entries in $\{-1,0,1\}$. In this case, the $\ell_{0}$-distance, the $\ell_{1}$-norm of all entries, and the Frobenius norm $\|B\|_{F}=\left(\sum_{i=1}^{m} \sum_{j=1}^{m} B_{i, j}^{2}\right)^{1 / 2}$ are all equivalent:

$$
\|B\|_{0}=\sum_{i=1}^{m} \sum_{j=1}^{m}\left|B_{i, j}\right|=\|B\|_{F}^{2} .
$$

Here, we also allow inner product functions mapping to numbers other than $\{0,1\}$ (note that no such number can match any entry in $A$ ).

The following problem is closely related to the Binary $\ell_{0}$-Rank- $k$ problem.
Hypercube Segmentation: Given a matrix $A \in\{0,1\}^{m \times n}$ and an integer $k$, we seek to compute vectors $u_{1}, \ldots, u_{k} \in\{0,1\}^{m}$ maximizing the value $\sum_{j=1}^{n} \min _{1 \leqslant \ell \leqslant k}\left(m-\left\|A_{:, j}-u_{\ell}\right\|_{0}\right)$ (or equivalently, minimizing the expression $\left.\sum_{j=1}^{n} \min _{1 \leqslant \ell \leqslant k}\left\|A_{:, j}-u_{\ell}\right\|_{0}\right)$.

The problem is NP-hard even for $k=2$ [Fei14]. Note that $m-\left\|A_{:, j}-u_{\ell}\right\|_{0}$ is the number of entries that column $A_{:, j}$ and vector $u_{\ell}$ have in common. It can be checked that the optimum is always at least $m n / 2$, and thus approximating the maximization version is easier, in the sense that any $\alpha$-approximation algorithm for minimization implies an $\alpha$-approximation for maximization. For the maximization version, Kleinberg et al. [KPR04] designed a polynomial-time 0.878 -approximation for any fixed $k$, which was improved to an efficient PTAS for any fixed $k$ by Alon and Sudakov [AS99]. For the minimization version, a PTAS was designed by Ostrovsky and Rabani [OR00]. This was extended to several clustering problems that are natural generalizations of Hypercube Segmentation [OR00, ABS10]. An efficient PTAS is known for a Euclidean variant of this clustering problem [BHI02].

Our results extend this line of work, since (the minimization version of) Hypercube Segmentation is a special case of the Generalized Binary $\ell_{0}$-Rank-k problem. Indeed, let $\pi:\{0,1\}^{k} \rightarrow\{1, \ldots, k\}$ be any onto function, e.g., $\pi$ maps the $i$-th vector in $\{0,1\}^{k}$ (w.r.t. any fixed ordering) to the number $\min \{i, k\}$. We define an inner product function $\langle x, y\rangle \stackrel{\text { def }}{=} x_{\pi(y)}$, i.e., $y$ specifies a coordinate of the vector $x$. Consider the matrix product $A^{\prime}=U \cdot V$ with respect to this inner product. Writing $\ell_{j} \stackrel{\text { def }}{=} \pi\left(V_{:, j}\right)$, we have $A_{i, j}^{\prime}=U_{i, \ell_{j}}$. In other words, the $j$-th column of $A^{\prime}$ is equal to one of the columns of $U$, and we can choose which one. Writing the columns of $U$ as $u_{1}, \ldots, u_{k}$ yields the equivalence with Hypercube Segmentation.

### 1.2 Binary $\ell_{0}$-rank-1 With Small Optimal Value

The results in this Section are proven in Chapter 3. Given a matrix $A \in\{0,1\}^{m \times n}$, our goal is to compute an approximate solution of the Binary $\ell_{0}$-Rank- 1 problem, and let us denote the optimal value by

$$
\begin{equation*}
\mathrm{OPT} \stackrel{\text { def }}{=} \min _{u \in\{0,1\}^{m}, v \in\{0,1\}^{n}}\left\|A-u \cdot v^{T}\right\|_{0} . \tag{1.1}
\end{equation*}
$$

In practice, approximating a matrix $A$ by a rank- 1 matrix $u v^{T}$ makes most sense if $A$ is close to being rank- 1 . Hence, the above optimization problem is most relevant in the case OPT $\ll\|A\|_{0}$. For this reason, we focus in this section on the case $\mathrm{OPT} /\|A\|_{0} \leqslant \phi$, for sufficiently small $\phi>0$.

We first give an algorithm that requires as an input a parameter $\phi \geqslant \mathrm{OPT} /\|A\|_{0}$.
Theorem 1.4. Given $A \in\{0,1\}^{m \times n}$ with row and column sums, and given $\phi \in\left(0, \frac{1}{80}\right]$ with $\mathrm{OPT} /\|A\|_{0} \leqslant$ $\phi$, we can compute in time $O\left(\min \left\{\|A\|_{0}+m+n, \phi^{-1}(m+n) \log (m n)\right\}\right)$ vectors $\widetilde{u} \in\{0,1\}^{m}$ and $\widetilde{v} \in\{0,1\}^{n}$ such that w.h.p. $\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0} \leqslant(1+5 \phi) \mathrm{OPT}+37 \phi^{2}\|A\|_{0}$.

Then, we use a $(2+\varepsilon)$-approximation algorithm for the Reals $\ell_{0}$-Rank- 1 problem that captures as a special case the Binary $\ell_{0}$-Rank-1 problem. We prove the following result in Chapter 4, see Section 4.2.

Theorem 1.5. Given $A \in\{0,1\}^{m \times n}$ with column adjacency arrays and OPT $\geqslant 1$, and given $\varepsilon \in(0,0.1]$, we can compute w.h.p. in time

$$
O\left(\left(\frac{n \log m}{\varepsilon^{2}}+\min \left\{\|A\|_{0}, n+\psi^{-1} \frac{\log n}{\varepsilon^{2}}\right\}\right) \frac{\log ^{2} n}{\varepsilon^{2}}\right)
$$

a column $A_{:, j}$ and a vector $z \in\{0,1\}^{n}$ such that w.h.p. $\left\|A-A_{:, j} \cdot z^{T}\right\|_{0} \leqslant(2+\varepsilon)$ OPT. Further, we can compute an estimate $Y$ such that w.h.p. $(1-\varepsilon) \mathrm{OPT} \leqslant Y \leqslant(2+2 \varepsilon)$ OPT.

Using Theorem 1.4 in combination with Theorem 1.5, we obtain a $(1+500 \psi)$-approximation algorithm for the Binary $\ell_{0}$-Rank-1 problem that does not need the parameter $\phi$ as an input.

Theorem 1.6. Given $A \in\{0,1\}^{m \times n}$ with column adjacency arrays and with row and column sums, for $\psi=\mathrm{OPT} /\|A\|_{0}$ we can compute w.h.p. in time $O\left(\min \left\{\|A\|_{0}+m+n, \psi^{-1}(m+n)\right\} \cdot \log ^{3}(m n)\right)$ vectors $\widetilde{u} \in\{0,1\}^{m}$ and $\widetilde{v} \in\{0,1\}^{n}$ such that w.h.p. $\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0} \leqslant(1+500 \psi) \mathrm{OPT}$.

Our algorithm simultaneously improves the approximation factor and the running time. Further, it even runs in sublinear $o\left(\|A\|_{0}\right)$ time, unlike all algorithms in previous works [SJY09, JPHY14]. Notably, when matrix $A$ is very well approximated by a low rank matrix, i.e. $\psi$ is a sub-constant, we obtain a $(1+o(1))$-approximation which is significantly better than the previous best known 2 -approximations.

Moreover, we also show that the running time of our algorithm is optimal up to a poly $\log (m n)$ factor, by proving that any $(1+O(\psi))$-approximation algorithm succeeding with constant probability must read $\Omega\left(\psi^{-1}(m+n)\right)$ entries of $A$ in the worst case.
Theorem 1.7. Let $C \geqslant 1$. Given an $n \times n$ binary matrix $A$ with column adjacency arrays and with row and column sums, and given $\sqrt{\log (n) / n} \ll \phi \leqslant 1 / 100 C$ such that $\mathrm{OPT} /\|A\|_{0} \leqslant \phi$, computing a $(1+C \phi)$-approximation of OPT requires to read $\Omega(n / \phi)$ entries of $A$ (in the worst case over $A$ ).

Furthermore, a variant of the algorithm from Theorem 1.6 can also be used to solve exactly the Binary $\ell_{0}$-Rank- 1 problem. This yields the following theorem, which in particular shows that the problem is in polynomial time when OPT $\leqslant O\left(\sqrt{\|A\|_{0}} \log (m n)\right)$.
Theorem 1.8. Given a matrix $A \in\{0,1\}^{m \times n}$, if $\mathrm{OPT} /\|A\|_{0} \leqslant 1 / 240$ then we can solve exactly the Binary $\ell_{0}$-Rank-1 problem in time $2^{O\left(\mathrm{OPT} / \sqrt{\|A\|_{0}}\right)} \cdot \operatorname{poly}(m n)$.

### 1.3 Algorithms for Reals $\ell_{0}$-Rank- $k$ Problem

The results in this Section are proven in Chapter 4. We establish approximation algorithms for several important variants of $\ell_{0}$-low rank approximation, which significantly improve the running time and the approximation factor of previous works. In some cases our algorithms even run in sublinear time, i.e., faster than reading all non-zero entries of the matrix. This is provably impossible for other measures such as the Frobenius norm and more generally, any $\ell_{p}$-norm for $p>0$. For $k>1$, our approximation algorithms are, to the best of our knowledge, the first with provable guarantees for these problems.

### 1.3.1 Preliminaries

For a matrix $A \in \mathbb{A}^{m \times n}$ with $m \geqslant n$ and entries $A_{i, j}$, let $A_{i,:}$ be its $i$-th row and $A_{:, j}$ be its $j$-th column. An algorithm that on input a sufficiently dense matrix $A$, runs in time $o\left(\|A\|_{0}\right)$ is called sublinear. In particular, we say that time $\widetilde{O}(m+n)$ is sublinear, since in general $m+n \ll\|A\|_{0}$. In this section, we study variants of the following problem.

Reals $\ell_{0}$-Rank- $k$ Given a matrix $A \in \mathbb{R}^{m \times n}$ and an integer $k$, compute matrices $U \in \mathbb{R}^{m \times k}$ and $V \in \mathbb{R}^{k \times n}$ minimizing $\|A-U \cdot V\|_{0}$. An algorithm for the Reals $\ell_{0}$-Rank- $k$ problem is an $\alpha$-approximation, if it outputs matrices $\widetilde{U} \in \mathbb{R}^{m \times k}$ and $\widetilde{V} \in \mathbb{R}^{k \times n}$ satisfying

$$
\|A-\widetilde{U} \cdot \widetilde{V}\|_{0} \leqslant \alpha \min _{U \in \mathbb{R}^{m \times k}, V \in \mathbb{R}^{k \times n}}\|A-U \cdot V\|_{0}
$$

Input Formats We always assume that we have random access to the entries of the given matrix $A$, i.e. we can read any entry $A_{i, j}$ in constant time. For our sublinear time algorithms we need more efficient access to the matrix, specifically the following two variants:
(1) We say that we are given $A$ with column adjacency arrays if we are given arrays $B_{1}, \ldots, B_{n}$ and lengths $\ell_{1}, \ldots, \ell_{n}$ such that for any $k \in\left\{1, \ldots, \ell_{j}\right\}$ the pair $B_{j}[k]=\left(i, A_{i, j}\right)$ stores the row $i$ containing the $k$-th nonzero entry in column $j$ as well as that entry $A_{i, j}$. This is a standard representation of matrices used in many applications. Note that given only these adjacency arrays $B_{1}, \ldots, B_{n}$, in order to access any entry $A_{i, j}$ we can perform a binary search over $B_{j}$, and hence random access to any matrix entry is in time $O(\log n)$. Moreover, we assume to have random access to matrix entries in constant time, and note that this is optimistic by at most a factor $O(\log n)$.
(2) We say that we are given matrix $A$ with row and column sums if we can access the numbers $\sum_{j} A_{i, j}$ for $i \in[m]$ and $\sum_{i} A_{i, j}$ for $j \in[n]$ in constant time (and, as always, access any entry $A_{i, j}$ in constant time). Notice that storing the row and column sums takes $O(m+n)$ space, and thus while this might not be standard information it is very cheap to store.

In Subsection 4.2.5, we show that the first access type even allows to sample from the set of nonzero entries uniformly in constant time.

Lemma 1.9. Given a matrix $A \in \mathbb{R}^{m \times n}$ with column adjacency arrays, after $O(n)$ time preprocessing we can sample a uniformly random nonzero entry $(i, j)$ from $A$ in time $O(1)$.

### 1.3.2 Reals $\ell_{0}$-rank- $k$

Given a matrix $A \in \mathbb{R}^{m \times n}$, our goal is to compute an approximate solution of the Reals $\ell_{0}$-rank- $k$ problem, and let us denote the optimal value by

$$
\begin{equation*}
\mathrm{OPT} \stackrel{\text { def }}{=} \min _{U \in \mathbb{R}^{m \times k}, V \in \mathbb{R}^{k \times n}}\|A-U \cdot V\|_{0} . \tag{1.2}
\end{equation*}
$$

We first give an impractical algorithm that runs in time $n^{O(k)}$ and achieves $O\left(k^{2}\right)$ factor approximation. To the best of our knowledge this is the first approximation algorithm for the Reals $\ell_{0}$-Rank- $k$ problem with non-trivial approximation guarantees.

Theorem 1.10. Given a matrix $A \in \mathbb{R}^{m \times n}$ and an integer $k$, we can compute in time $O\left(n^{k+1} m^{2} k^{\omega+1}\right)$ a set of $k$ indices $J^{(k)} \subset[n]$ and a matrix $Z \in \mathbb{R}^{k \times n}$ such that $\left\|A-A_{:, J^{(k)}} \cdot Z\right\|_{0} \leqslant O\left(k^{2}\right) \cdot \mathrm{OPT}$.

To make our algorithm practical, we reduce the running time to poly $(m n)$, with an exponent independent of $k$, if we allow for a bicriteria solution. In particular, we allow the algorithm to output a matrix $A^{\prime}$ of larger rank $O(k \log (n / k))$ and approximation factor $O\left(k^{2} \log (n / k)\right)$.

Theorem 1.11. Given a matrix $A \in \mathbb{R}^{m \times n}$ and an integer $k$, there is an algorithm that in expected time $\operatorname{poly}(m, n)$ outputs a subset of indices $J \subset[n]$ with $|J|=O(k \log (n / k))$ and a matrix $Z \in \mathbb{R}^{|J| \times n}$ such that $\left\|A-A_{;, J} \cdot Z\right\|_{0} \leqslant O\left(k^{2} \log (n / k)\right) \cdot$ OPT.

Although, we do not obtain exactly rank $k$, many of the motivations for finding a low rank approximation, such as reducing the number of parameters and fast matrix-vector multiplication, still hold if the output rank is $O(k \log (n / k))$. We are not aware of any alternative algorithms that achieve poly $(m n)$ time and any provable approximation factor, even for bicriteria solutions.

### 1.3.3 Reals $\ell_{0}$-rank-1

Given a rank-1 matrix $A \in \mathbb{R}^{m \times n}$, there is an algorithm that runs in time $O\left(\|A\|_{0}\right)$ and finds the exact rank-1 decomposition $u v^{T}$ of $A$. Here, we focus on the case when $A$ is not a rank-1 matrix.

The previous best algorithm due to Jiang et al. [JPHY14] was based on the observation that there exists a column $u$ of $A$ spanning a 2 -approximation. Hence, solving the problem $\min _{v}\left\|A-u \cdot v^{T}\right\|_{0}$ for each column $u$ of matrix $A$ yields a 2-approximation. The problem $\min _{v \in \mathbb{R}^{n}}\left\|A-u \cdot v^{T}\right\|_{0}$ decomposes into $\sum_{i} \min _{i}\left\|A_{:, i}-v_{i} u\right\|_{0}$, where $A_{:, i}$ is the $i$-th column of $A$, and $v_{i}$ the $i$-th entry of vector $v$. The optimal $v_{i}$ is the mode of the ratios $A_{i, j} / u_{j}$, where $j$ ranges over indices in $\{1,2, \ldots, m\}$ with $u_{j} \neq 0$. As a result, one can find a rank-1 matrix $u v^{T}$ providing a 2-approximation in time $O\left(n\|A\|_{0}\right)$, which was the best known running time.

We design randomized algorithms for solving this problem. Let the optimal value be

$$
\begin{equation*}
\mathrm{OPT} \stackrel{\text { def }}{=} \min _{u \in \mathbb{R}^{m}, v \in \mathbb{R}^{n}}\left\|A-u \cdot v^{T}\right\|_{0} \tag{1.3}
\end{equation*}
$$

Our algorithm yields a $(2+\varepsilon)$-approximation and runs in nearly linear time in $\|A\|_{0}$, for any constant $\varepsilon>0$. Moreover, a variant of our algorithm even runs in sublinear time when OPT $\geqslant\left(\varepsilon^{-1} \log (m n)\right)^{4}$ and $\|A\|_{0} \geqslant n\left(\varepsilon^{-1} \log (m n)\right)^{4}$.

Theorem 1.12. Given $A \in \mathbb{R}^{m \times n}$ with column adjacency arrays and $\mathrm{OPT} \geqslant 1$, and given $\varepsilon \in(0,0.1]$, there is an algorithm that runs w.h.p. in time

$$
O\left(\left(\frac{n \log m}{\varepsilon^{2}}+\min \left\{\|A\|_{0}, n+\psi^{-1} \frac{\log n}{\varepsilon^{2}}\right\}\right) \frac{\log ^{2} n}{\varepsilon^{2}}\right) \quad \text { where } \quad \psi=\frac{\mathrm{OPT}}{\|A\|_{0}}
$$

and outputs a column $A_{:, j}$ and a vector $z \in \mathbb{R}^{n}$ such that w.h.p. $\left\|A-A_{:, j} \cdot z^{T}\right\|_{0} \leqslant(2+\varepsilon)$ OPT. The algorithm also computes an estimate $Y$ satisfying w.h.p. $(1-\varepsilon) \mathrm{OPT} \leqslant Y \leqslant(2+2 \varepsilon) \mathrm{OPT}$.

This significantly improves upon the earlier $O\left(n\|A\|_{0}\right)$ time algorithm for not too small $\varepsilon$ and $\psi$. Our result should be contrasted to Frobenius norm low rank approximation, for which $\Omega\left(\|A\|_{0}\right)$ time is required even for $k=1$, as otherwise one might miss a very large entry in $A$. Since $\ell_{0}$-low rank approximation is insensitive to the magnitude of entries of $A$, we bypass this general impossibility result.

## Chapter 2

## A PTAS For Generalized Binary $\ell_{0}$-Rank- $k$

Given a matrix $A \in\{0,1\}^{m \times n}$ with $m \geqslant n$, an integer $k$, and an inner product function $\langle.,\rangle:.\{0,1\}^{k} \times$ $\{0,1\}^{k} \rightarrow \mathbb{R}$, the Generalized Binary $\ell_{0}$-Rank- $k$ problem asks to find matrices $U \in\{0,1\}^{m \times k}$ and $V \in\{0,1\}^{k \times n}$ minimizing $\|A-U \cdot V\|_{0}$, where the product $U \cdot V$ is the $m \times n$ matrix $B$ with $B_{i, j}=\left\langle U_{i,:}, V_{:, j}\right\rangle$. An algorithm for the Generalized Binary $\ell_{0}$-Rank- $k$ problem is an $\alpha$-approximation, if it outputs matrices $U \in\{0,1\}^{m \times k}$ and $V \in\{0,1\}^{k \times n}$ satisfying

$$
\|A-U \cdot V\|_{0} \leqslant \alpha \cdot \min _{U^{\prime} \in\{0,1\}^{m \times k}, V^{\prime} \in\{0,1\}^{k \times n}}\left\|A-U^{\prime} \cdot V^{\prime}\right\|_{0} .
$$

As shown in Chapter 1, by choosing an appropriate inner product function $\langle.,$.$\rangle which also runs in time$ $O(k)$, we obtain the Binary $\ell_{0}$-Rank- $k$ problem over the reals, $\mathbb{F}_{2}$, and the Boolean semiring. We assume that the function $\langle.,$.$\rangle can be evaluated in time 2^{O(k)}$, in order to simplify our running time bounds.

### 2.1 Technical Overview

### 2.1.1 PTAS For Generalized Binary $\ell_{0}$-Rank- $k$

The Generalized Binary $\ell_{0}$-Rank- $k$ problem can be rephrased as a clustering problem with constrained centers, whose goal is to choose a set of centers satisfying a certain system of linear equations, in order to minimize the total $\ell_{0}$-distance of all columns of $A$ to their closest center. The main difference to usual clustering problems is that the centers cannot be chosen independently.

We view the choice of matrix $U$ as picking a set of "cluster centers" $S_{U} \stackrel{\text { def }}{=}\left\{U \cdot y \mid y \in\{0,1\}^{k}\right\}$. Observe that any column of $U \cdot V$ is in $S_{U}$, and thus we view the choice of column $V_{:, j}$ as picking one of the constrained centers in $S_{U}$. Formally, we rephrase the Generalized Binary $\ell_{0}$-Rank- $k$ problem as

$$
\begin{align*}
\min _{U \in\{0,1\}^{m \times k}, V \in\{0,1\}^{k \times n}}\|A-U \cdot V\|_{0} & =\min _{U \in\{0,1\}^{m \times k}} \sum_{j=1}^{n} \min _{V_{:, j} \in\{0,1\}^{k}}\left\|A_{:, j}-U \cdot V_{:, j}\right\|_{0} \\
& =\min _{U \in\{0,1\}^{m \times k}} \sum_{j=1}^{n} \min _{s \in S_{U}}\left\|A_{:, j}-s\right\|_{0} . \tag{2.1}
\end{align*}
$$

Any matrix $V$ gives rise to a "clustering" as partitioning $C_{V}=\left(C_{y}\right)_{y \in\{0,1\}^{k}}$ of the columns of $V$ with $C_{y}=\left\{j \in[n] \mid V_{:, j}=y\right\}$. If we knew an optimal clustering $C=C_{V}$, for some optimal matrix $V$, we could compute an optimal matrix $U$ as the best response to $V$. Note that

$$
\min _{U \in\{0,1\}^{m \times k}} \sum_{y \in\{0,1\}^{k}} \sum_{j \in C_{y}}\left\|A_{:, j}-U \cdot y\right\|_{0}=\sum_{i=1}^{m} \min _{U_{i,:} \in\{0,1\}^{k}} \sum_{y \in\{0,1\}^{k}} \sum_{j \in C_{y}}\left\|A_{i, j}-U_{i,:} \cdot y\right\|_{0} .
$$

Therefore, given $C$ we can compute independently for each $i \in[m]$ the optimal row $U_{i,:} \in\{0,1\}^{k}$, by enumerating over all possible binary vectors of dimension $k$ and selecting the one that minimizes the summation $\sum_{y \in\{0,1\}^{k}} \sum_{j \in C_{y}}\left\|A_{i, j}-U_{i,:} \cdot y\right\|_{0}$.

What if instead we could only sample from $C$ ? That is, suppose that we are allowed to draw a constant number $t=\operatorname{poly}\left(2^{k} / \varepsilon\right)$ of samples from each of the optimal clusters $C_{y}$ uniformly at random. Denote by $\widetilde{C}_{y}$ the samples drawn from $C_{y}$. A natural approach is to replace the exact cost above by the following unbiased estimator:

$$
\widetilde{E} \stackrel{\text { def }}{=} \sum_{y \in\{0,1\}^{k}} \frac{\left|C_{y}\right|}{\left|\widetilde{C}_{y}\right|} \cdot \sum_{j \in \widetilde{C}_{y}}\left\|A_{:, j}-U \cdot y\right\|_{0} .
$$

We show that with good probability any matrix $U=U(\widetilde{C})$ minimizing the estimated cost $\widetilde{E}$ is close to an optimal solution. In particular, we prove for any matrix $V \in\{0,1\}^{k \times n}$ that

$$
\begin{equation*}
\mathbb{E}_{\widetilde{C}}\left[\|A-U(\widetilde{C}) \cdot V\|_{0}\right] \leqslant(1+\varepsilon) \cdot \min _{U \in\{0,1\}^{m \times k}}\|A-U \cdot V\|_{0} \tag{2.2}
\end{equation*}
$$

The biggest issue in proving statement (2.2) is that the number of samples $t=\operatorname{poly}\left(2^{k} / \varepsilon\right)$ is independent of the ambient space dimension $n$. A key prior probabilistic result, established by Alon and Sudakov [AS99], gives an additive $\pm \varepsilon m n$ approximation for the maximization version of a clustering problem with unconstrained centers, known as Hypercube Segmentation. Since the optimum value of this maximization problem is always at least $m n / 2$, a multiplicative factor $(1+\varepsilon)$-approximation is obtained. Our contribution is twofold. First, we generalize their analysis to clustering problems with constrained centers, and second we prove a multiplicative factor $(1+\varepsilon)$-approximation for the minimization version. The proof of (2.2) takes a significant fraction of this chapter.

We combine the sampling result (2.2) with the following observations to obtain a deterministic polynomial time approximation scheme (PTAS) in time $m \cdot n^{\text {poly }\left(2^{k} / \varepsilon\right)}$. We later discuss how to further improve this running time. Let $U, V$ be an optimal solution to the Generalized Binary $\ell_{0}$-Rank- $k$ problem.
(1) To evaluate the estimated cost $\widetilde{E}$, we need the sizes $\left|C_{y}\right|$ of an optimal clustering $C$. We can guess these sizes with an $n^{2^{k}}$ overhead in the running time. In fact, it suffices to know these cardinalities approximately, see Lemma 2.5, and thus this overhead ${ }^{1}$ can be reduced to $\left(t+\varepsilon^{-1} \cdot \log n\right)^{2^{k}}$.
(2) Using the (approximate) size $\left|C_{y}\right|$ and the samples $\widetilde{C}_{y}$ drawn u.a.r. from $C_{y}$, for all $y \in\{0,1\}^{k}$, we can compute in time $2^{O(k)} m n$ a matrix $U(\widetilde{C})$ minimizing the estimated cost $\widetilde{E}$, since the estimator $\widetilde{E}$ can be split into a sum over the rows of $U(\widetilde{C})$ and each row is chosen independently as a minimizer among all possible binary vectors of dimension $k$.
(3) Given $U(\widetilde{C})$, we can compute a best response matrix $V(\widetilde{C})$ which has cost $\|A-U(\widetilde{C}) \cdot V(\widetilde{C})\|_{0} \leqslant$ $\|A-U(\widetilde{C}) \cdot V\|_{0}$, and thus by $(2.2)$ the expected cost at most $(1+\varepsilon) \mathrm{OPT}$.
(4) The only remaining step is to draw samples $\widetilde{C}_{y}$ from the optimal clustering. However, in time $O\left(n^{2^{k} t}\right)=n^{\text {poly }\left(2^{k} / \varepsilon\right)}$ we can enumerate all possible families $\left(\widetilde{C}_{y}\right)_{y \in\{0,1\}^{k}}$, and the best such family yields a solution that is at least as good as a random sample. In total, we obtain a PTAS in time $m \cdot n^{\text {poly }\left(2^{k} / \varepsilon\right)}$.

The largest part of this chapter is devoted to make the above PTAS efficient, i.e., to reduce the running time from $m \cdot n^{\text {poly }\left(2^{k} / \varepsilon\right)}$ to $(2 / \varepsilon)^{2^{O(k)} / \varepsilon^{2}} \cdot m n^{1+o(1)}$, where $o(1)$ hides a factor $(\log \log n)^{1.1} / \log n$. By the preceding outline, it suffices to speed up Steps (1) and (4), i.e., to design a fast algorithm that guesses approximate cluster sizes and samples from the optimal clusters.

The standard sampling approach for clustering problems such as $k$-means [KSS04] is as follows. At least one of the clusters of the optimal solution is "large", say $\left|C_{y}\right| \geqslant n / 2^{k}$. Sample $t$ columns uniformly at random from the set $[n]$ of all columns. Then with probability at least $\left(1 / 2^{k}\right)^{t}$ all samples lie in $C_{y}$, and in this case they form a uniform sample from this cluster. In the usual situation without restrictions on the cluster centers, the samples from $C_{y}$ allow us to determine an approximate cluster center $\widetilde{s}^{(y)}$. Do this as long as large clusters exist (recall that we have guessed approximate cluster sizes in Step (1), so we know which clusters are large). When all remaining clusters are small, remove the $n / 2$ columns that are closest to the approximate cluster centers $\widetilde{s}^{(y)}$ determined so far, and estimate the cost of these columns using the centers $\widetilde{s}^{(y)}$. As there are no restrictions on the cluster centers, this yields a good cost estimation of the removed columns, and since the $\ell_{0}$-distance is additive the algorithm recurses on the remaining columns, i.e. on an instance of twice smaller size. We continue this process until each cluster is sampled. This approach has been used to obtain linear time approximation schemes for $k$-means and $k$-median in a variety of ambient spaces [KSS04, KSS05, ABS10].

The issue in our situation is that we cannot fix a cluster center $\widetilde{s}^{(y)}$ by looking only at the samples $\widetilde{C}_{y}$, since we have dependencies among cluster centers. We nevertheless make this approach work, by showing that a uniformly random column $r^{(y)} \in[n]$ is a good "representative" of the cluster $C_{y}$ with not-too-small probability. In the case when all remaining clusters are small, we then simply remove the $n / 2$ columns that are closest to the representatives $r^{(y)}$ of the clusters that we already sampled from. Although these

[^1]representatives can be far from the optimal cluster centers due to the linear restrictions on the latter, we show in Section 2.3 that nevertheless this algorithm yields samples from the optimal clusters.

We prove that the preceding algorithm succeeds with probability at least $(\varepsilon / t)^{2^{O(k) \cdot t}}$. Further, we show that the approximate cluster sizes $\left|\widetilde{C}_{y}\right|$ of an optimal clustering can be guessed with an overhead of $\left(2^{k} / \varepsilon\right)^{2^{O(k)}} \cdot(\log n)^{(\log \log n)^{0.1}}$. In contrast to the standard clustering approach, the representatives $r^{(y)}$ do not yield a good cost estimation of the removed columns. We overcome this issue by first collecting all samples $\widetilde{C}$ from the optimal clusters, and then computing approximate cluster centers that satisfy certain linear constraints, i.e. a matrix $U(\widetilde{C})$ and its best response matrix $V(\widetilde{C})$. The latter computation runs in linear time $2^{O(k)} \cdot m n$ in the size of the original instance, and this in combination with the guessing overhead, yields the total running time of $(2 / \varepsilon)^{2^{O(k)} / \varepsilon^{2}} \cdot m n^{1+o(1)}$. For further details, we refer the reader to Algorithm 2 in Subsection 2.3.2.

Our algorithm achieves a substantial generalization of the standard clustering approach and applies to the situation with constrained centers. This yields the first randomized almost-linear time approximation scheme (PTAS) for the Generalized Binary $\ell_{0}$-Rank- $k$ problem.
Theorem 1.1 (from page 5). (PTAS) For any error $\varepsilon \in(0,1 / 2)$, there is a $(1+\varepsilon)$-approximation algorithm for the Generalized Binary $\ell_{0}-R a n k-k$ problem that runs in time $(2 / \varepsilon)^{2^{O(k)} / \varepsilon^{2}} \cdot m n^{1+o(1)}$ and succeeds with constant probability ${ }^{2}$, where o(1) hides a factor $(\log \log n)^{1.1} / \log n$.

Our running time is close to optimal, in the sense that the running time of any PTAS for the Generalized Binary $\ell_{0}$-Rank- $k$ problem must depend exponentially on $1 / \varepsilon$ and doubly exponentially on $k$, assuming the Exponential Time Hypothesis (ETH). We show this in the following.

Theorem 1.2 (from page 5). (Hardness for Generalized Binary $\ell_{0}$-Rank-k) Assuming the Exponential Time Hypothesis, the Generalized Binary $\ell_{0}-$ Rank- $k$ problem has no $(1+\varepsilon)$-approximation algorithm in time $2^{1 / \varepsilon^{o(1)}} \cdot 2^{m^{o(1)}}$. Further, for any $\varepsilon \geqslant 0$, the Generalized Binary $\ell_{0}-$ Rank-k problem has no $(1+\varepsilon)-$ approximation algorithm in time $2^{2^{o(k)}} \cdot 2^{m^{o(1)}}$.

Regarding the dependence on $\varepsilon$, assume w.l.o.g. that $m \geqslant n$, and thus the input size is $O(m)$. Even for $k=1$ the problem is known to be NP-hard [GV15, DAJ ${ }^{+}$15]. Under ETH, no NP-hard problem has an algorithm ${ }^{3}$ in time $2^{m^{o(1)}}$. We can restrict to $\varepsilon \geqslant 1 / m^{2}$, since any better approximation already yields an optimal solution. It follows for $k=1$ that the Generalized Binary $\ell_{0}$-Rank- $k$ problem has no $(1+\varepsilon)$-approximation algorithm in time $2^{1 / \varepsilon^{o(1)}} \cdot 2^{m^{o(1)}}$. In other words, in order to improve our exponential dependence on $1 / \varepsilon$ to subexponential, we would need to pay an exponential factor in $m$.

Regarding the dependence on $k$, note that for any $\varepsilon$ a $(1+\varepsilon)$-approximation algorithm for our problem decides whether the answer is 0 or larger. In particular, over the Boolean semiring it solves the problem whether a given bipartite graph can be covered with $k$ bicliques. For this problem, Chandran et al. [CIK16] proved that even for $k=O(\log m)$ there is no algorithm in time $2^{2^{o(k)}}$, unless ETH fails. It follows that for any $\varepsilon \geqslant 0$, Generalized Binary $\ell_{0}$-Rank- $k$ has no $(1+\varepsilon)$-approximation algorithm in time $2^{2^{o(k)}} \cdot 2^{o(m)}$. In other words, in order to improve our doubly exponential dependence on $k$, we would need to pay an exponential factor in $m$. Together, this shows that our running time is close to optimal.

Organization In Subsection 2.2.2, we state our core sampling result. In Subsection 2.2.3, we give a simple but inefficient deterministic PTAS for the Generalized Binary $\ell_{0}$-Rank- $k$ problem, which serves as a blueprint for our efficient randomized PTAS. We present first the deterministic PTAS as it is conceptually simple and exhibits the main algorithmic challenge, namely, to design an efficient sampling procedure. In Subsection 2.2.4, we prove our core sampling result by extending the analysis of Alon and Sudakov [AS99] to clustering problems with constrained centers, and by further strengthening an additive $\pm \varepsilon m n$ approximation guarantee to a multiplicative factor $(1+\varepsilon)$-approximation. In Subsection 2.3, we design an efficient sampling procedure, and this yields our efficient randomized PTAS. Our approach uses ideas from clustering algorithms pioneered by Kumar et al. [KSS04] and refined in [KSS05, ABS10].

In Section 2.4, we give a faster randomized PTAS for the Binary $\ell_{0}$-Rank- 1 problem which improves the algorithm in Theorem 1.1 and runs in time $O\left((1 / \varepsilon)^{1 / \varepsilon^{2}} \cdot\left(\|A\|_{0}+m+n\right) \cdot \log ^{3}(m n)\right)$.

[^2]
### 2.2 Constant Size Sampling Suffices

### 2.2.1 Preliminaries

Chebyshev's inequality We now give some basic facts. Let $Z_{1}, \ldots, Z_{n}$ be independent Bernoulli random variables, with $Z_{i} \sim \operatorname{Ber}\left(p_{i}\right)$. Let $Z \stackrel{\text { def }}{=} Z_{1}+\ldots+Z_{n}$ and $\mu \stackrel{\text { def }}{=} \mathbb{E}[Z]$.
Lemma 2.1. For any $\Delta>0$, we have $\operatorname{Pr}[|Z-\mu|>\Delta] \leqslant \mu / \Delta^{2}$.
Proof. By independence, we have

$$
\operatorname{Var}(Z)=\sum_{i=1}^{n} \operatorname{Var}\left(Z_{i}\right)=\sum_{i=1}^{n} p_{i}\left(1-p_{i}\right) \leqslant \sum_{i=1}^{n} p_{i}=\mu .
$$

By Chebyshev's inequality, for any $\Delta>0$ it holds that $\operatorname{Pr}[|Z-\mu|>\Delta] \leqslant \operatorname{Var}(Z) / \Delta^{2}$. The claim follows by $\operatorname{Var}(Z) \leqslant \mu$.

Lemma 2.2. For any $\Delta>0$, we have $\operatorname{Pr}[|Z-\mu|>\Delta] \leqslant \sqrt{n} / \Delta$.
Proof. As in the previous lemma's proof, we have $\operatorname{Pr}[|Z-\mu|>\Delta] \leqslant \operatorname{Var}(Z) / \Delta^{2}$, where $\operatorname{Var}(Z) \leqslant \mu \leqslant n$, and thus $\operatorname{Pr}[|Z-\mu|>\Delta] \leqslant n / \Delta^{2}$. It also follows that $\operatorname{Pr}[|Z-\mu|>\Delta] \leqslant \sqrt{n} / \Delta$, since if $\sqrt{n} / \Delta<1$ we have $n / \Delta^{2} \leqslant \sqrt{n} / \Delta$, and otherwise the inequality is trivial.

### 2.2.2 Sampling Theorem

We denote the optimal value of Generalized Binary $\ell_{0}$-Rank- $k$ by

$$
\mathrm{OPT}=\mathrm{OPT}_{k} \stackrel{\text { def }}{=} \min _{U \in\{0,1\}^{m \times k}, V \in\{0,1\}^{k \times n}}\|A-U \cdot V\|_{0} .
$$

Further, for a fixed matrix $V \in\{0,1\}^{k \times n}$ we let

$$
\mathrm{OPT}_{k}^{V} \stackrel{\text { def }}{=} \min _{U \in\{0,1\}^{m \times k}}\|A-U \cdot V\|_{0}
$$

and we say that a matrix $U \in\{0,1\}^{m \times k}$ is a best response to $V$, if $\|A-U \cdot V\|_{0}=\mathrm{OPT}_{k}^{V}$.
Given $A \in\{0,1\}^{m \times n}$, an integer $k$, and an inner product $\langle.,\rangle:.\{0,1\}^{k} \times\{0,1\}^{k} \rightarrow \mathbb{R}$, let $V \in\{0,1\}^{k \times n}$ be arbitrary and $U \in\{0,1\}^{m \times k}$ be a best response to $V$, i.e.,

$$
\mathrm{OPT}_{k}^{V} \stackrel{\text { def }}{=}\|A-U \cdot V\|_{0}=\min _{U^{\prime} \in\{0,1\}^{m \times k}}\left\|A-U^{\prime} \cdot V\right\|_{0}
$$

Partition the columns of $V$ (equivalently the columns of $A$ ) into sets

$$
C_{y}^{V} \stackrel{\text { def }}{=}\left\{j: V_{:, j}=y\right\}
$$

for all $y \in\{0,1\}^{k}$. For any row $i$, vector $y \in\{0,1\}^{k}$, and $c \in\{0,1\}$ we define by

$$
Z_{i, y, c} \stackrel{\text { def }}{=}\left|\left\{j \in C_{y}^{V} \mid A_{i j}=c\right\}\right| \quad \text { and } \quad Z_{i, y, \neq c} \stackrel{\text { def }}{=}\left|\left\{j \in C_{y}^{V} \mid A_{i j} \neq c\right\}\right| .
$$

Then, the exact cost of a row $i$ for any vector $x \in\{0,1\}^{k}$ is given by

$$
\begin{equation*}
E_{i, x} \stackrel{\text { def }}{=}\left\|A_{i,:}-x^{T} \cdot V\right\|_{0}=\sum_{y \in\{0,1\}^{k}} Z_{i, y, \neq\langle x, y\rangle} . \tag{2.3}
\end{equation*}
$$

Observe that $U_{i,:} \in\{0,1\}^{k}$ is a vector $x$ minimizing $E_{i, x}$ (this follows from $U$ being a best response to $V)$, and let $E_{i} \stackrel{\text { def }}{=} E_{i, U_{i,}}$.

We do not know the partitioning $\left\{C_{y}^{V}\right\}_{y \in\{0,1\}^{k}}$, however, as we will see later we can assume that (1) we can sample elements from each $C_{y}^{V}$ and (2) we have good approximations of the sizes $\left|C_{y}^{V}\right|$ for all $y$.

For (1), to set up notation let $\widetilde{C}=\left(\widetilde{C}_{y}\right)_{y \in\{0,1\}^{k}}$ be a family, where $\widetilde{C}_{y}$ is a random multiset with elements from $C_{y}^{V}$. Specifically, we will work with the following distribution $\mathcal{D}_{V, t}$ for some $t \in \mathbb{N}$ : For any
$y \in\{0,1\}^{k}$, if $\left|C_{y}^{V}\right|<t$ let $\widetilde{C}_{y}=C_{y}^{V}$, otherwise sample $t$ elements from $C_{y}^{V}$ with replacement and let the resulting multiset be $\widetilde{C}_{y}$.

For (2), we say that a sequence $\alpha=\left(\alpha_{y}\right)_{y \in\{0,1\}^{k}}$ is a sequence of $\delta$-approximate cluster sizes if for any $y \in\{0,1\}^{k}$ with $\left|C_{y}^{V}\right|<t$ we have $\alpha_{y}=\left|C_{y}^{V}\right|$, and for the remaining $y \in\{0,1\}^{k}$ we have

$$
\left|C_{y}^{V}\right| \leqslant \alpha_{y} \leqslant(1+\delta)\left|C_{y}^{V}\right|
$$

Then corresponding to $Z_{i, y, c}$ and $Z_{i, y, \neq c}$ we have random variables

$$
\widetilde{Z}_{i, y, c} \stackrel{\text { def }}{=}\left|\left\{j \in \widetilde{C}_{y}: A_{i, j}=c\right\}\right| \quad \text { and } \quad \widetilde{Z}_{i, y, \neq c} \stackrel{\text { def }}{=}\left|\left\{j \in \widetilde{C}_{y}: A_{i, j} \neq c\right\}\right| .
$$

Given $\widetilde{C}$ and $\alpha$, we define the estimated cost of row $i$ and vector $x \in\{0,1\}^{k}$ as

$$
\begin{equation*}
\widetilde{E}_{i, x} \stackrel{\text { def }}{=} \sum_{y \in\{0,1\}^{k}} \frac{\alpha_{y}}{\left|\widetilde{C}_{y}\right|} \widetilde{Z}_{i, y, \neq\langle x, y\rangle} \tag{2.4}
\end{equation*}
$$

If $C_{y}^{V}=\emptyset$ for some $y \in\{0,1\}^{k}$, then $\widetilde{Z}_{i, y, \neq\langle x, y\rangle}=0$ and we define the corresponding summand in (2.4) to be 0 . Observe that if the approximation $\alpha_{y}$ is exact, i.e., $\alpha_{y}=\left|C_{y}^{V}\right|$, then $\widetilde{E}_{i, x}$ is an unbiased estimator for the exact cost $E_{i, x}$.

We now simplify the problem to optimizing the estimated cost instead of the exact cost. Specifically, we construct a matrix $\widetilde{U} \in\{0,1\}^{m \times k}$ by picking for each row $i$ any

$$
\widetilde{U}_{i,:} \in \operatorname{argmin}\left\{\widetilde{E}_{i, x}: x \in\{0,1\}^{k}\right\} .
$$

Note that matrix $\widetilde{U}$ depends on the input $(A, k,\langle.,\rangle$.$) , on the sequence \alpha$, and on the sampled multisets $\widetilde{C}=\left(\widetilde{C}_{y}\right)_{y \in\{0,1\}^{k}}$. When it is clear from the context, we suppress the dependence on $A, k,\langle.,$.$\rangle , and write$ $\widetilde{U}=\widetilde{U}(\widetilde{C}, \alpha)$. We generalize now (2.2), and show that this matrix yields a good approximation of the optimal cost.

Theorem 2.3. For any matrix $V \in\{0,1\}^{k \times n}$, let $\alpha$ be a sequence of $\frac{\varepsilon}{6}$-approximate cluster sizes and draw $\widetilde{C}$ according to distribution $\mathcal{D}_{V, t}$ for $t=t(k, \varepsilon) \stackrel{\text { def }}{=} 2^{4 k+14} / \varepsilon^{2}$. Then we have

$$
\mathbb{E}_{\widetilde{C}}\left[\|A-\widetilde{U}(\widetilde{C}, \alpha) \cdot V\|_{0}\right] \leqslant(1+\varepsilon) \mathrm{OPT}_{k}^{V}
$$

We defer the proof of Theorem 2.3 to Section 2.2.4, and first show how it yields a simple but inefficient deterministic PTAS for the Generalized Binary $\ell_{0}$-Rank- $k$ problem running in time $m \cdot n^{\operatorname{poly}\left(2^{k} / \varepsilon\right)}$, see Section 2.2.3. Then, in Section 2.3, we design a sampling procedure that improves the running time to $(2 / \varepsilon)^{2^{O(k)}} / \varepsilon^{2} \cdot m n^{1+o(1)}$, where $o(1)$ hides a factor $(\log \log n)^{1.1} / \log n$.

### 2.2.3 Simple PTAS

In this subsection, we show how Theorem 2.3 leads to a simple but inefficient deterministic PTAS, see Algorithm 1, for the Generalized Binary $\ell_{0}$-Rank- $k$ problem.

A basic, but crucial property used in our analysis is that given a matrix $A \in\{0,1\}^{m \times n}$, an integer $k$ and a matrix $V$, we can compute a best response matrix $U$ minimizing $\|A-U \cdot V\|_{0}$ in time $2^{O(k)} m n$. Indeed, we can split $\|A-U \cdot V\|_{0}=\sum_{i=1}^{n}\left\|A_{i,:}-U_{i,:} \cdot V\right\|_{0}$ and brute-force the optimal solution $U_{i,:} \in\{0,1\}^{k}$ minimizing the $i$-th summand $\left\|A_{i,:}-U_{i,:} \cdot V\right\|_{0}$. Symmetrically, given $U$ we can compute a best response $V$ in time $2^{O(k)} m n$. In particular, if $(U, V)$ is an optimal solution then $U$ is a best response for $V$, and $V$ is a best response for $U$.

We now present the pseudocode of Algorithm 1.

```
Algorithm 1 Simple deterministic PTAS for the Generalized Binary \(\ell_{0}\)-Rank- \(k\) problem
Input: A matrix \(A \in\{0,1\}^{m \times n}\), an integer \(k\), an inner product \(\langle.,\).\(\rangle , and \varepsilon \in(0,1)\).
Output: Matrices \(\widetilde{U} \in\{0,1\}^{m \times k}, \widetilde{V} \in\{0,1\}^{k \times n}\) such that \(\|A-\widetilde{U} \cdot \widetilde{V}\|_{0} \leqslant(1+\varepsilon) \mathrm{OPT}_{k}\).
```

1. (Guess column set sizes) Let $U, V$ be an optimal solution. Exhaustively guess all sizes $\left|C_{y}^{V}\right|=: \alpha_{y}$ for $y \in\{0,1\}^{k}$. There are $n^{2^{k}}$ possibilities.
2. (Guess column multisets) Theorem 2.3 implies existence of a family $\widetilde{C}=\left(\widetilde{C}_{y}\right)_{y \in\{0,1\}^{k}}$ such that $\|A-\widetilde{U}(\widetilde{C}, \alpha) \cdot V\|_{0} \leqslant(1+\varepsilon) \mathrm{OPT}_{k}$, where each $\widetilde{C}_{y}$ is a multiset consisting of at most $t$ indices in $\{1, \ldots, n\}$. Exhaustively guess such a family $\widetilde{C}$. There are $n^{O\left(t \cdot 2^{k}\right)}$ possibilities.
3. (Compute $\widetilde{U}$ ) Now we know $A, k,\langle.,\rangle,.\left|C_{y}^{V}\right|$ for all $y \in\{0,1\}^{k}$, and $\widetilde{C}$, thus we can compute the matrix $\widetilde{U}=\widetilde{U}(\widetilde{C}, \alpha)$, where row $\widetilde{U}_{i, \text { : }}$ is any vector $x$ minimizing the estimated cost $\widetilde{E}_{i, x}$. Since each row $\widetilde{U}_{i,:} \in\{0,1\}^{k}$ can be optimized independently, this takes time $2^{O(k)} m n$. If we guessed correctly, we have $\|A-\widetilde{U} \cdot V\|_{0} \leqslant(1+\varepsilon) \mathrm{OPT}_{k}$.
4. (Compute $\widetilde{V}$ ) Compute $\widetilde{V}$ as a best response to $\widetilde{U}$. This takes time $2^{O(k)} m n$. If we guessed correctly, by best-response and Step 3, we have

$$
\|A-\widetilde{U} \cdot \tilde{V}\|_{0} \leqslant\|A-\widetilde{U} \cdot V\|_{0} \leqslant(1+\varepsilon) \mathrm{OPT}_{k}
$$

5. Return the pair $(\widetilde{U}, \widetilde{V})$ minimizing $\|A-\widetilde{U} \cdot \widetilde{V}\|_{0}$ over all exhaustive guesses.

The correctness of Algorithm 1 immediately follows from Theorem 2.3. The running time is dominated by the exhaustive guessing in Step 2 , so we obtain time $m \cdot n^{\text {poly }\left(2^{k} / \varepsilon\right)}$.

### 2.2.4 Proof of the Sampling Theorem 2.3

We follow the notation in Section 2.2.2, in particular $V \in\{0,1\}^{k \times n}$ is an arbitrary matrix and $U \in$ $\{0,1\}^{m \times k}$ is a best response to $V$. We define $D_{i, x}$ as the difference of the cost of row $i$ w.r.t. a vector $x$ and the cost of row $i$ w.r.t. the optimal vector $U_{i, \text { : }}$, i.e.,

$$
\begin{align*}
D_{i, x} \stackrel{\text { def }}{=} E_{i, x}-E_{i} & =\left\|A_{i,:}-x^{T} \cdot V\right\|_{0}-\left\|A_{i,:}-U_{i,:} \cdot V\right\|_{0}  \tag{2.5}\\
& =\sum_{y \in\{0,1\}^{k}} Z_{i, y, \neq\langle x, y\rangle}-Z_{i, y, \neq\left\langle U_{i,:}, y\right\rangle} .
\end{align*}
$$

Note that a vector $x$ is suboptimal for a row $i$ if and only if $D_{i, x}>0$. By a straightforward splitting of the expectation, we obtain the following.

Claim 2.4. For every $V \in\{0,1\}^{k \times n}$, we have

$$
\mathbb{E}_{\widetilde{C}}\left[\|A-\widetilde{U} \cdot V\|_{0}\right]=\mathrm{OPT}_{k}^{V}+\sum_{i=1}^{m} \sum_{\substack{x \in\{0,1\}^{k} \\ D_{i, x}>0}} \operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \cdot D_{i, x}
$$

Proof. We split $\|A-\widetilde{U} \cdot V\|_{0}=\sum_{i=1}^{m}\left\|A_{i,:}-\widetilde{U}_{i,:} \cdot V\right\|_{0}$. This yields

$$
\begin{aligned}
\mathbb{E}_{\widetilde{C}}\left[\|A-\widetilde{U} \cdot V\|_{0}\right] & =\sum_{i=1}^{m} \mathbb{E}_{\widetilde{C}}\left[\left\|A_{i,:}-\widetilde{U}_{i,:} \cdot V\right\|_{0}\right] \\
& =\sum_{i=1}^{m} \sum_{x \in\{0,1\}^{k}} \operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \cdot\left\|A_{i,:}-x^{T} \cdot V\right\|_{0}
\end{aligned}
$$

By definition of $D_{i, x}$, we have

$$
\begin{aligned}
\mathbb{E}_{\widetilde{C}}\left[\|A-\widetilde{U} \cdot V\|_{0}\right] & =\sum_{i=1}^{m} \sum_{x \in\{0,1\}^{k}} \operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \cdot\left(\left\|A_{i,:}-U_{i,:} \cdot V\right\|_{0}+D_{i, x}\right) \\
& =\sum_{i=1}^{m}\left(\left\|A_{i,:}-U_{i,:} \cdot V\right\|_{0}+\sum_{x \in\{0,1\}^{k}} \operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \cdot D_{i, x}\right) \\
& =\operatorname{OPT}_{k}^{V}+\sum_{i=1}^{m} \sum_{x \in\{0,1\}^{k}} \operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \cdot D_{i, x} \\
& =\operatorname{OPT}_{k}^{V}+\sum_{i=1}^{m} \sum_{\substack{x \in\{0,1\}^{k} \\
D_{i, x}>0}} \operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \cdot D_{i, x} .
\end{aligned}
$$

Similarly to $D_{i, x}$, we define an estimator

$$
\begin{equation*}
\widetilde{D}_{i, x} \stackrel{\text { def }}{=} \widetilde{E}_{i, x}-\widetilde{E}_{i, U_{i,:}}=\sum_{y \in\{0,1\}^{k}} \frac{\alpha_{y}}{\left|\widetilde{C}_{y}\right|} \cdot\left(\widetilde{Z}_{i, y, \neq\langle x, y\rangle}-\widetilde{Z}_{i, y, \neq\left\langle U_{i,:}, y\right\rangle}\right) . \tag{2.6}
\end{equation*}
$$

Note that $\widetilde{U}_{i, \text { : }}$ is chosen among the vectors $x \in\{0,1\}^{k}$ minimizing $\widetilde{D}_{i, x}$. Hence, our goal is to show that significantly suboptimal vectors (with $D_{i, x}>\frac{\varepsilon}{3} \cdot E_{i}$ ) satisfy $\widetilde{D}_{i, x}>0$ with good probability, and thus these vectors are not picked in $\widetilde{U}$.

To this end, we split the rows $i$ and suboptimal vectors $x$ into:

$$
\begin{aligned}
& L_{0} \stackrel{\text { def }}{=}\left\{(i, x): 0<D_{i, x} \leqslant \frac{\varepsilon}{3} \cdot E_{i}\right\}, \\
& L_{1} \stackrel{\text { def }}{=}\left\{(i, x): \frac{\varepsilon}{3} \cdot E_{i}<D_{i, x} \leqslant E_{i}\right\}, \\
& L_{2} \stackrel{\text { def }}{=}\left\{(i, x): D_{i, x}>E_{i}\right\} .
\end{aligned}
$$

Observe that $\sum_{(x, i) \in L_{0}} \operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \cdot D_{i, x} \leqslant \frac{\varepsilon}{3} \cdot \mathrm{OPT}_{k}^{V}$. By Claim 2.4, we can ignore all tuples $(i, x) \in L_{0}$, since

$$
\begin{equation*}
\mathbb{E}_{\widetilde{C}}\left[\|A-\widetilde{U} \cdot V\|_{0}\right] \leqslant\left(1+\frac{\varepsilon}{3}\right) \mathrm{OPT}_{k}^{V}+\sum_{(i, x) \in L_{1} \cup L_{2}} \operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \cdot D_{i, x} \tag{2.7}
\end{equation*}
$$

Hence, our goal is to upper bound the summation $\sum_{(i, x) \in L_{1} \cup L_{2}} \operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \cdot D_{i, x}$.
We next establish a sufficient condition for $\widetilde{U}_{i, \text { : }} \neq x$, for any suboptimal vector $x$. Note that by definition of $D_{i, x}$ we have

$$
\begin{equation*}
D_{i, x}=\sum_{y \in\{0,1\}^{k}} Z_{i, y, \neq\langle x, y\rangle}-Z_{i, y, \neq\left\langle U_{i,:}, y\right\rangle}=\sum_{y \in \hat{Y}_{i, x}} Z_{i, y, \neq\langle x, y\rangle}-Z_{i, y, \neq\left\langle U_{i,:}, y\right\rangle} \tag{2.8}
\end{equation*}
$$

where $\hat{Y}_{i, x} \stackrel{\text { def }}{=}\left\{y \in\{0,1\}^{k}:\langle x, y\rangle \neq\left\langle U_{i,:}, y\right\rangle\right\}$. Similarly, for the estimator we have

$$
\begin{equation*}
\widetilde{D}_{i, x}=\sum_{y \in\{0,1\}^{k}} \frac{\alpha_{y}}{\left|\widetilde{C}_{y}\right|} \cdot\left(\widetilde{Z}_{i, y, \neq\langle x, y\rangle}-\widetilde{Z}_{i, y, \neq\left\langle U_{i,:}, y\right\rangle}\right)=\sum_{y \in \hat{Y}_{i, x}} \frac{\alpha_{y}}{\left|\widetilde{C}_{y}\right|} \cdot\left(\widetilde{Z}_{i, y, \neq\langle x, y\rangle}-\widetilde{Z}_{i, y, \neq\left\langle U_{i,:}, y\right\rangle}\right) . \tag{2.9}
\end{equation*}
$$

Let $\mathcal{W}_{i, x}$ be the event that for every $y \in Y_{i, x} \stackrel{\text { def }}{=}\left\{y \in \hat{Y}_{i, x}:\left|\widetilde{C}_{y}\right|=t\right\}$ and every $c \in\{0,1\}$, we have

$$
\left|\widetilde{Z}_{i, y, c}-\frac{\left|\widetilde{C}_{y}\right|}{\left|C_{y}^{V}\right|} \cdot Z_{i, y, c}\right| \leqslant \Delta_{y}, \quad \text { where } \quad \Delta_{y} \stackrel{\text { def }}{=} \frac{t \cdot D_{i, x}}{2^{k+2} \cdot \alpha_{y}}
$$

We now show that conditioned on the event $\mathcal{W}_{i, x}$, we have $\widetilde{D}_{i, x}>0$ for any $(i, x) \in L_{1} \cup L_{2}$, and thus $\widetilde{U}_{i,:} \neq x$.
Lemma 2.5. For any vector $x \in\{0,1\}^{k}$ and row $i \in[m]$, if event $\mathcal{W}_{\sim}{\underset{\sim}{D}}^{\prime}$ occurs then it follows that $\widetilde{D}_{i, x} \geqslant \frac{1}{2} \cdot D_{i, x}-\frac{\varepsilon}{6} \cdot E_{i}$. In particular, if additionally $D_{i, x}>\frac{\varepsilon}{3} \cdot E_{i}$ then $\widetilde{D}_{i, x}>0$.

Proof. Observe that $\widetilde{Z}_{i, y, \neq c} \in\left\{\widetilde{Z}_{i, y, 0}, \widetilde{Z}_{i, y, 1}, \widetilde{Z}_{i, y, 0}+\widetilde{Z}_{i, y, 1}\right\}$ for any $i, y, c$. Since

$$
\mathbb{E}\left[\widetilde{Z}_{i, y, 0}+\widetilde{Z}_{i, y, 1}\right]=\left|\widetilde{C}_{y}\right|=\widetilde{Z}_{i, y, 0}+\widetilde{Z}_{i, y, 1},
$$

conditioned on the event $\mathcal{W}_{i, x}$, for any $y \in Y_{i, x}$, all three random variables $\widetilde{Z}_{i, y, 0}, \widetilde{Z}_{i, y, 1}$ and $\widetilde{Z}_{i, y, 0}+\widetilde{Z}_{i, y, 1}$ differ from their expectation by at most $\Delta_{y}$. Hence, we have

$$
\left|\frac{\alpha_{y}}{\widetilde{C}_{y} \mid} \widetilde{Z}_{i, y, \neq\langle x, y\rangle}-\frac{\alpha_{y}}{\left|C_{y}^{V}\right|} Z_{i, y, \neq\langle x, y\rangle}\right| \leqslant \frac{D_{i, x}}{2^{k+2}} .
$$

The same inequality also holds for $y \in \hat{Y}_{i, x} \backslash Y_{i, x}$, since then $\widetilde{Z}_{i, y, \neq\langle x, y\rangle}=Z_{i, y, \neq\langle x, y\rangle}$ and $\left|\widetilde{C}_{y}\right|=\left|C_{y}^{V}\right|$ (by definition of the distribution $\mathcal{D}_{V, t}$. In combination with (2.9) we obtain

$$
\begin{equation*}
\widetilde{D}_{i, x} \geqslant-\frac{D_{i, x}}{2}+\sum_{y \in \hat{X}_{i, x}} \frac{\alpha_{y}}{\left|C_{y}^{V}\right|} \cdot\left(Z_{i, y, \neq \neq x, y\rangle}-Z_{i, y, \neq\left\langle U_{i,, s, y\rangle}\right.}\right) . \tag{2.10}
\end{equation*}
$$

Let $\alpha_{y}=\left(1+\gamma_{y}\right)\left|C_{y}^{V}\right|$ with $0 \leqslant \gamma_{y} \leqslant \frac{\varepsilon}{6}$ for any $y \in\{0,1\}^{k}$. By (2.8), and since $\alpha_{y}=\left|C_{y}^{V}\right|=\left|\widetilde{C}_{y}\right|$ for every $y \in \hat{Y}_{i, x} \backslash Y_{i, x}$ (by definition of distribution $\mathcal{D}_{V, t}$ ), we have

$$
\begin{aligned}
\sum_{y \in \hat{Y}_{i, x}} \frac{\alpha_{y}}{\left|C_{y}^{V}\right|}\left(Z_{i, y, \neq\langle\langle x, y\rangle}-Z_{i, y, \neq\left\langle U_{i, i}, y\right\rangle}\right) & =D_{i, x}+\sum_{y \in Y_{i, x}} \gamma_{y}\left(Z_{i, y, \neq\langle\langle x, y\rangle}-Z_{i, y, \neq\left\langle U U_{i, i}, y\right\rangle}\right) \\
& \geqslant D_{i, x}-\sum_{y \in Y_{i, x}} \gamma_{y} Z_{i, y, \neq\left\langle U_{i,, y}, y\right\rangle} \\
& \geqslant D_{i, x}-\frac{\varepsilon}{6} \sum_{y \in\{0,1\}^{k}} Z_{i, y, \neq\left\langle\left\langle U_{i, i}, y\right\rangle\right.}=D_{i, x}-\frac{\varepsilon}{6} E_{i} .
\end{aligned}
$$

Together with (2.10), we have $\widetilde{D}_{i, x} \geqslant \frac{1}{2} D_{i, x}-\frac{\varepsilon}{6} E_{i}$.
We next upper bound the probability of picking a suboptimal vector $x$.
Claim 2.6. For any $x \in\{0,1\}^{k}$ with $D_{i, x}>\frac{\varepsilon}{3} \cdot E_{i}$, we have

$$
\operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \leqslant \sum_{y \in Y_{i, x}} \min _{c \in\{0,1\}} \operatorname{Pr}\left[\left|\widetilde{Z}_{i, y, c}-\mathbb{E}\left[\widetilde{Z}_{i, y, c}\right]\right|>\Delta_{y}\right] .
$$

Proof. For any $y \in\{0,1\}^{k}$, we have

$$
\widetilde{Z}_{i, y, 0}+\widetilde{Z}_{i, y, 1}=\left|\widetilde{C}_{y}\right|=\mathbb{E}\left[\widetilde{Z}_{i, y, 0}\right]+\mathbb{E}\left[\widetilde{Z}_{i, y, 1}\right] .
$$

Further, it holds that

$$
\left|\widetilde{Z}_{i, y, 0}-\mathbb{E}\left[\widetilde{Z}_{i, y, 0}\right]\right|=\left|\widetilde{Z}_{i, y, 1}-\mathbb{E}\left[\widetilde{Z}_{i, y, 1}\right]\right|,
$$

and thus

$$
\begin{aligned}
& \operatorname{Pr}\left[\left|\widetilde{Z}_{i, y, 0}-\mathbb{E}\left[\widetilde{Z}_{i, y, 0}\right]\right| \leqslant \Delta_{y}\right]=\operatorname{Pr}\left[\left|\widetilde{Z}_{i, y, 1}-\mathbb{E}\left[\widetilde{Z}_{i, y, 1}\right]\right| \leqslant \Delta_{y}\right] \\
& =\operatorname{Pr}\left[\left|\widetilde{Z}_{i, y, 0}-\mathbb{E}\left[\widetilde{Z}_{i, y, 0}\right]\right| \leqslant \Delta_{y} \text { and }\left|\widetilde{Z}_{i, y, 1}-\mathbb{E}\left[\widetilde{Z}_{i, y, 1}\right]\right| \leqslant \Delta_{y}\right] .
\end{aligned}
$$

Since $\widetilde{U}_{i,:}=x$ can only hold if $\widetilde{D}_{i, x} \leqslant 0$, the claim follows by Lemma 2.5 and a union bound over $y \in Y_{i, x}$.

In the following subsections, we bound the summation in (2.7) over the sets $L_{1}$ and $L_{2}$.

## Case 1: Small Difference

We show first that $\left|L_{1}\right|$ is small (see Claim 2.7). Then, we use a simple bound for $\operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right]$ which is based on Lemma 2.2 (see Claim 2.8).

Claim 2.7. It holds that

$$
\sum_{(i, x) \in L_{1}} \sum_{y \in Y_{i, x}}\left|C_{y}^{V}\right| \leqslant 2^{k+2} \cdot \mathrm{OPT}_{k}^{V} .
$$

Proof. Fix $(i, x) \in L_{1}$ and let $y \in Y_{i, x}$. Note that since $\langle x, y\rangle \neq\left\langle U_{i,:}, y\right\rangle$ we have

$$
\left\{j \in C_{y}^{V}: A_{i, j} \neq\langle x, y\rangle\right\} \cup\left\{j \in C_{y}^{V}: A_{i, j} \neq\left\langle U_{i,:}, y\right\rangle\right\}=C_{y}^{V}
$$

Note that this union is not necessarily disjoint, e.g., if $\langle x, y\rangle \notin\{0,1\}$. Since $E_{i, x}=D_{i, x}+E_{i}$ (by (2.5)) and $D_{i, x} \leqslant E_{i}$ (by definition of $L_{1}$ ), we have

$$
\begin{equation*}
\sum_{y \in Y_{i, x}}\left|C_{y}^{V}\right| \leqslant \sum_{y \in Y_{i, x}} Z_{i, y, \neq\langle x, y\rangle}+Z_{i, y, \neq\left\langle U_{i, r}, y\right\rangle} \leqslant E_{i, x}+E_{i} \leqslant 3 E_{i} . \tag{2.11}
\end{equation*}
$$

Fixing $x$ and summing over all $i$ with $(i, x) \in L_{1}$, the term $E_{i}$ sums to at most $\mathrm{OPT}_{k}^{V}$. Also summing over all $x \in\{0,1\}^{k}$ yields another factor $2^{k}$. Therefore, the claim follows.

Claim 2.8. It holds that

$$
\sum_{(i, x) \in L_{1}} \operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \cdot D_{i, x} \leqslant \frac{\varepsilon}{3} \cdot \mathrm{OPT}_{k}^{V}
$$

Proof. Note that for any row $i$, vector $y \in Y_{i, x}$, and $c \in\{0,1\}$, the random variable $\widetilde{Z}_{i, y, c}$ is a sum of independent Bernoulli random variables, since the $t$ samples from $C_{y}^{V}$ forming $\widetilde{C}_{y}$ are independent, and each sample contributes either 0 or 1 to $\widetilde{Z}_{i, y, c}$. Hence, our instantiations of Chebyshev's inequality, Lemmas 2.1 and 2.2, are applicable. We use Lemma 2.2 to bound

$$
\operatorname{Pr}\left[\left|\widetilde{Z}_{i, y, c}-\mathbb{E}\left[\widetilde{Z}_{i, y, c}\right]\right|>\Delta_{y}\right] \leqslant \frac{\sqrt{t}}{\Delta_{y}}
$$

Since $\Delta_{y}=t \cdot D_{i, x} /\left(2^{k+2} \cdot \alpha_{y}\right)$ and $\alpha_{y} \leqslant\left(1+\frac{\varepsilon}{6}\right)\left|C_{y}^{V}\right|<2\left|C_{y}^{V}\right|$, we have

$$
\operatorname{Pr}\left[\left|\widetilde{Z}_{i, y, c}-\mathbb{E}\left[\widetilde{Z}_{i, y, c}\right]\right|>\Delta_{y}\right] \leqslant \frac{2^{k+3}\left|C_{y}^{V}\right|}{\sqrt{t} \cdot D_{i, x}}
$$

and thus by Claim 2.6, we obtain

$$
\operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \leqslant \frac{2^{k+3}}{\sqrt{t} \cdot D_{i, x}} \sum_{y \in Y_{i, x}}\left|C_{y}^{V}\right| .
$$

Claim 2.7 now yields

$$
\sum_{(i, x) \in L_{1}} \operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \cdot D_{i, x} \leqslant \frac{2^{k+3}}{\sqrt{t}} \sum_{(i, x) \in L_{1}} \sum_{y \in Y_{i, x}}\left|C_{y}^{V}\right| \leqslant \frac{2^{2 k+5}}{\sqrt{t}} \mathrm{OPT}_{k}^{V}
$$

Since we chose $t \geqslant 2^{4 k+14} / \varepsilon^{2}$, see Theorem 2.3, we obtain the upper bound $\frac{\varepsilon}{3} \mathrm{OPT}_{k}^{V}$.

## Case 2: Large Difference

We use here the stronger instantiation of Chebyshev's inequality, Lemma 2.1, and charge $\mu=\mathbb{E}\left[\widetilde{Z}_{i, y, c}\right]$ against $\mathrm{OPT}_{k}^{V}$.

Claim 2.9. It holds that

$$
\sum_{(i, x) \in L_{2}} \operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \cdot D_{i, x} \leqslant \frac{\varepsilon}{3} \cdot \mathrm{OPT}_{k}^{V}
$$

Proof. Fix $(i, x) \in L_{2}$ and let $y \in Y_{i, x}$. As in the proof of Claim 2.8, we see that our instantiation of Chebyshev's inequality, Lemma 2.1, is applicable to $\widetilde{Z}_{i, y, c}$ for any $c \in\{0,1\}$. We obtain

$$
\operatorname{Pr}\left[\left|\widetilde{Z}_{i, y, c}-\mathbb{E}\left[\widetilde{Z}_{i, y, c}\right]\right|>\Delta_{y}\right] \leqslant \frac{\mathbb{E}\left[\widetilde{Z}_{i, y, c}\right]}{\Delta_{y}^{2}}
$$

Note that $\mathbb{E}\left[\widetilde{Z}_{i, y, c}\right]=Z_{i, y, c} \cdot t /\left|C_{y}^{V}\right|$, since $\left|\widetilde{C}_{y}\right|=t$. Using $\min _{c \in\{0,1\}} Z_{i, y, c} \leqslant Z_{i, y, \neq\left\langle U_{i,:}, y\right\rangle}$, we have

$$
\min _{c \in\{0,1\}} \operatorname{Pr}\left[\left|\widetilde{Z}_{i, y, c}-\mathbb{E}\left[\widetilde{Z}_{i, y, c}\right]\right|>\Delta_{y}\right] \leqslant \frac{t}{\left|C_{y}^{V}\right| \Delta_{y}^{2}} \cdot Z_{i, y, \neq\left\langle U_{i,:}, y\right\rangle}
$$

Since $\Delta_{y}=t \cdot D_{i, x} /\left(2^{k+2} \cdot \alpha_{y}\right)$ and $\alpha_{y} \leqslant(1+\varepsilon / 6)\left|C_{y}^{V}\right|<2\left|C_{y}^{V}\right|$, we have

$$
\min _{c \in\{0,1\}} \operatorname{Pr}\left[\left|\widetilde{Z}_{i, y, c}-\mathbb{E}\left[\widetilde{Z}_{i, y, c}\right]\right|>\Delta_{y}\right] \leqslant \frac{2^{2 k+6} \cdot\left|C_{y}^{V}\right|}{t \cdot\left(D_{i, x}\right)^{2}} \cdot Z_{i, y, \neq\left\langle U_{i,:}, y\right\rangle}
$$

Summing over all $y \in Y_{i, x}$, Claim 2.6 yields

$$
\begin{equation*}
\operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \leqslant \sum_{y \in Y_{i, x}} \frac{2^{2 k+6} \cdot\left|C_{y}^{V}\right|}{t \cdot\left(D_{i, x}\right)^{2}} \cdot Z_{i, y, \neq\left\langle U_{i,:}, y\right\rangle} \tag{2.12}
\end{equation*}
$$

We again use inequality (2.11), i.e., $\sum_{y \in Y_{i, x}}\left|C_{y}^{V}\right| \leqslant E_{i, x}+E_{i}$. Since $E_{i, x}=D_{i, x}+E_{i}$ (by (2.5)) and $E_{i}<D_{i, x}$ (by definition of $L_{2}$ ), we have $\left|C_{y}^{V}\right| \leqslant 3 D_{i, x}$ for any $y \in Y_{i, x}$. Together with (2.12), and then using the definition of $E_{i}$, we have

$$
\operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \cdot D_{i, x} \leqslant \frac{2^{2 k+8}}{t} \sum_{y \in Y_{i, x}} Z_{i, y, \neq\left\langle U_{i,:}, y\right\rangle} \leqslant \frac{2^{2 k+8}}{t} E_{i},
$$

Fixing $x$ and summing over all $i$ with $(i, x) \in L_{2}$, the term $E_{i}$ sums to at most $\mathrm{OPT}_{k}^{V}$. Also summing over all $x \in\{0,1\}^{k}$ yields another factor $2^{k}$. Thus, it follows that

$$
\sum_{(i, x) \in L_{2}} \operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \cdot D_{i, x} \leqslant \frac{2^{3 k+8}}{t} \mathrm{OPT}_{k}^{V}
$$

Since we chose $t \geqslant 2^{3 k+10} / \varepsilon$, see Theorem 2.3, we obtain the upper bound $\frac{\varepsilon}{3} \cdot \mathrm{OPT}_{k}^{V}$.

### 2.2.5 Finishing the Proof

Taken together, the above claims prove Theorem 2.3.
Proof of Theorem 2.3. By Claim 2.4, splitting into $L_{0}, L_{1}$ and $L_{2}$, and using Claims 2.8 and 2.9, it follows for any $\varepsilon \in(0,1)$ and

$$
\begin{equation*}
t=2^{4 k+12} / \varepsilon^{2} \tag{2.13}
\end{equation*}
$$

that the expected approximate solution satisfies

$$
\begin{aligned}
\mathbb{E}_{\widetilde{V}}\left[\|A-\widetilde{U} \cdot V\|_{0}\right] & \leqslant\left(1+\frac{\varepsilon}{3}\right) \mathrm{OPT}_{k}^{V}+\sum_{(i, x) \in L_{1}} \operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \cdot D_{i, x}+\sum_{(i, x) \in L_{2}} \operatorname{Pr}\left[\widetilde{U}_{i,:}=x\right] \cdot D_{i, x} \\
& \leqslant(1+\varepsilon) \mathrm{OPT}_{k}^{V} .
\end{aligned}
$$

### 2.3 Efficient Sampling Algorithm

The conceptually simple PTAS in Section 2.2.3 has two running time bottlenecks, due to the exhaustive enumeration in Step 1 and Step 2. Namely, Step 1 guesses exactly the sizes $\left|C_{y}^{V}\right|$ for each $y \in\{0,1\}^{k}$, and there are $n^{O\left(2^{k}\right)}$ possibilities; and Step 2 guesses among all columns of matrix $A$ the multiset family $\widetilde{C}$, guaranteed to exist by Theorem 2.3, and there are $n^{O\left(t \cdot 2^{k}\right)}$ possibilities.

Since Theorem 2.3 needs only approximate cluster sizes, it suffices in Step 1 to guess numbers $\alpha_{y}$ with $\left|C_{y}^{V}\right| \leqslant \alpha_{y} \leqslant\left(1+\frac{\varepsilon}{6}\right)\left|C_{y}^{V}\right|$ if $\left|C_{y}^{V}\right| \geqslant t$, and $\alpha_{y}=\left|C_{y}^{V}\right|$ otherwise, where $t=2^{4 k+12} / \varepsilon^{2}$. Hence, the runtime overhead for Step 1 can be easily improved to $\left(t+\varepsilon^{-1} \log n\right)^{2^{k}}$.

To reduce the exhaustive enumeration in Step 2, we design an efficient sampling procedure, see Algorithm 2, that uses ideas from clustering algorithms pioneered by Kumar et al. [KSS04] and refined in [KSS05, ABS10]. Our algorithm reduces the total exhaustive enumeration in Step 2 and the guessing overhead for the approximate cluster sizes in Step 1 to $\left(2^{k} / \varepsilon\right)^{2^{O(k)}} \cdot(\log n)^{(\log \log n)^{0.1}}$ possibilities.

This section is structured as follows. We first replace an optimal solution $(U, V)$ by a "well-clusterable" solution ( $U, W$ ), which will help in our correctness proof. In Subsection 2.3.2 we present pseudocode for our sampling algorithm. We then prove its correctness in Subsection 2.3.3 and analyze its running time in Subsection 2.3.4. Finally, we show how to use the sampling algorithm designed in Subsection 2.3.2 together with the ideas of the simple PTAS from Subsection 2.2.3 to prove Theorem 1.1, see Subsection 2.3.5.

### 2.3.1 Existence of a $(U, V, \varepsilon)$-Clusterable Solution

For a matrix $B \in\{0,1\}^{m \times n}$ we denote by $\operatorname{ColSupp}(B)$ the set of unique columns of $B$. Note that if the columns of $U$ are linearly independent then $U \cdot \operatorname{ColSupp}(V)$ denotes the set of distinct columns of $U \cdot V$. In the clustering formulation of the Generalized Binary $\ell_{0}$-Rank- $k$ problem, as discussed in the introduction (2.1), the set $U \cdot \operatorname{ColSupp}(V)$ corresponds to the set of cluster centers.

Given matrices $U, V$, we will first replace $V$ by a related matrix $W$ in a way that makes all centers of $U \cdot \operatorname{ColSupp}(W)$ sufficiently different without increasing the cost too much, as formalized in the following.

Lemma 2.10. For any $U \in\{0,1\}^{m \times k}, V \in\{0,1\}^{k \times n}$ and $\varepsilon \in(0,1)$, there exists a matrix $W \in\{0,1\}^{k \times n}$ such that $\|A-U \cdot W\|_{0} \leqslant(1+\varepsilon)\|A-U \cdot V\|_{0}$ and for any distinct $y, z \in \operatorname{ColSupp}(W)$ we have
(i) $\|U \cdot y-U \cdot z\|_{0}>\varepsilon \cdot 2^{-k} \cdot\|A-U \cdot V\|_{0} / \min \left\{\left|C_{y}^{W}\right|,\left|C_{z}^{W}\right|\right\}$, and
(ii) $\left\|A_{:, j}-U \cdot y\right\|_{0} \leqslant\left\|A_{:, j}-U \cdot z\right\|_{0}$ for every $j \in C_{y}^{W}$.
$W e$ say that such a matrix $W$ is $(U, V, \varepsilon)$-clusterable.
Proof. The proof is by construction of $W$. We initialize $W \stackrel{\text { def }}{=} V$ and then iteratively resolve violations of (i) and (ii). In each step, resulting in a matrix $W^{\prime}$, we ensure that $\operatorname{ColSupp}\left(W^{\prime}\right) \subseteq \operatorname{ColSupp}(W)$. We call this support-monotonicity.

We can resolve all violations of (ii) at once by iterating over all columns $j \in[n]$ and replacing $W_{:, j}$ by the vector $z \in \operatorname{ColSupp}(W)$ minimizing $\left\|A_{:, j}-U \cdot z\right\|_{0}$. This does not increase the cost $\|A-U \cdot W\|_{0}$ and results in a matrix $W^{\prime}$ without any violations of (ii).

So assume that there is a violation of (i). That is, for distinct $y, z \in \operatorname{ColSupp}(W)$, where we can assume without loss of generality that $\left|C_{y}^{W}\right| \leqslant\left|C_{z}^{W}\right|$, we have $\|U \cdot y-U \cdot z\|_{0} \leqslant \varepsilon \cdot 2^{-k} \cdot\|A-U \cdot V\|_{0} /\left|C_{y}^{W}\right|$. We change the matrix $W$ by replacing for every $j \in C_{y}^{W}$ the column $W_{:, j}=y$ by $z$. Call the resulting matrix $W^{\prime}$. Note that the cost of any replaced column $j$ changes to

$$
\begin{aligned}
\left\|A_{:, j}-U \cdot W_{:, j}^{\prime}\right\|_{0}=\left\|A_{:, j}-U \cdot z\right\|_{0} & \leqslant\left\|A_{:, j}-U \cdot y\right\|_{0}+\|U \cdot y-U \cdot z\|_{0} \\
& \leqslant\left\|A_{:, j}-U \cdot W_{:, j}\right\|_{0}+\varepsilon \cdot 2^{-k} \cdot\|A-U \cdot V\|_{0} /\left|C_{y}^{W}\right|
\end{aligned}
$$

Since the number of replaced columns is $\left|C_{y}^{W}\right|$, the overall cost increase is at most $\varepsilon \cdot 2^{-k} \cdot\|A-U \cdot V\|_{0}$. Note that after this step the size of $\operatorname{ColSupp}(W)$ is reduced by 1 , since we removed any occurrence of column $y$. By support-monotonicity, the number of such steps is bounded by $2^{k}$. Since resolving violations of (ii) does not increase the cost, the final cost is bounded by $(1+\varepsilon)\|A-U \cdot V\|_{0}$.

After at most $2^{k}$ times resolving a violation of (i) and then all violations of (ii), we end up with a matrix $W$ without violations and the claimed cost bound.

### 2.3.2 The Algorithm Sample

Given $A \in\{0,1\}^{m \times n}, k \in \mathbb{N}, \varepsilon \in(0,1)$, and $t \in \mathbb{N}$, fix any optimal solution $U, V$, i.e., $\|A-U \cdot V\|_{0}=\mathrm{OPT}_{k}$. Our proof will use the additional structure provided by well-clusterable solutions. Therefore, fix any $(U, V, \varepsilon)$-clusterable matrix $W$ as in Lemma 2.10. Since $\|A-U \cdot W\|_{0} \leqslant(1+\varepsilon)\|A-U \cdot V\|_{0}$, we can restrict to matrix $W$. Specifically, we fix the optimal partitioning $C^{W}$ of $[n]$ for the purpose of the analysis and for the guessing steps of the algorithm. Our goal is to sample from the distribution $\mathcal{D}_{W, t}$.

Pseudocode of our sampling algorithm Sample ${ }_{A, k, \varepsilon, t}(M, \mathcal{N}, \widetilde{R}, \widetilde{C}, \alpha)$ is given below. The arguments of this procedure are as follows. Matrix $M$ is the current submatrix of $A$ (initialized as the full matrix $A$ ). Set $\mathcal{N} \subseteq\{0,1\}^{k}$ is the set of clusters that we did not yet sample from (initialized to $\{0,1\}^{k}$ ). The sequence $\widetilde{R}$ stores "representatives" of the clusters that we already sampled from (initialized to undefined entries $(\perp, \ldots, \perp))$. The sequence $\widetilde{C}$ contains our samples, so in the end we want $\widetilde{C}$ to be drawn according to $\mathcal{D}_{W, t}\left(\widetilde{C}\right.$ is initialized such that $\widetilde{C}_{y}=\emptyset$ for all $\left.y \in\{0,1\}^{k}\right)$. Finally, $\alpha$ contains guesses for the sizes of the clusters that we already sampled from, so in the end we want it to be a sequence of $\frac{\varepsilon}{6}$-approximate cluster sizes $\left(\alpha\right.$ is initialized such that $\alpha_{y}=0$ for all $\left.y \in\{0,1\}^{k}\right)$. This algorithm is closely related to algorithm "Irred- $k$-means" by Kumar et al. [KSS04], see the introduction for a discussion.

In this algorithm, at the base case we call EstimateBestResponse ${ }_{A, k}(\widetilde{C}, \alpha)$, which computes matrix $\widetilde{U}=\widetilde{U}(\widetilde{C}, \alpha)$ and a best response $\widetilde{V}$ to $\widetilde{U}$. Apart from the base case, there are three phases of algorithm Sample. In the sampling phase, we first guess some $y \in \mathcal{N}$ and an approximation $\alpha_{y}$ of $\left|C_{y}^{W}\right|$. Then we sample $\min \left\{t, \alpha_{y}\right\}$ columns from $M$ to form a multiset $\widetilde{C}_{y}$, and we sample one column from $\widetilde{R}^{\widetilde{R}}$ to form $\widetilde{R}_{y}$. We make a recursive call with $y$ removed from $\mathcal{N}$ and updated $\widetilde{R}, \widetilde{C}, \alpha$ by the values $\widetilde{R}_{y}, \widetilde{C}_{y}, \alpha_{y}$. As an intermediate solution, we let $U^{(1)}, V^{(1)}$ be the best solution returned by the recursive calls over all exhaustive guesses. In the pruning phase, we delete the $n_{M} / 2$ columns of $M$ that are closest to $\widetilde{R}$, and we make a recursive call with the resulting matrix $M^{\prime}$, not changing the remaining arguments. Denote the returned solution by $U^{(2)}, V^{(2)}$. Finally, in the decision phase we return the better solution between $U^{(1)}, V^{(1)}$ and $U^{(2)}, V^{(2)}$.

```
Algorithm 2 Efficient Sampling
\(\operatorname{Sample}_{A, k, \varepsilon, t}(M, \mathcal{N}, \widetilde{R}, \widetilde{C}, \alpha)\)
    let \(n_{M}\) be the number of columns of \(M\)
    set \(\nu \stackrel{\text { def }}{=}\left(\varepsilon / 2^{k+4}\right)^{2^{k}+2-|\mathcal{N}|}\)
1. If \(\mathcal{N}=\emptyset\) or \(n_{M}=0: \operatorname{Return}(\widetilde{U}, \widetilde{V})=\) EstimateBestResponse \(_{A, k}(\widetilde{C}, \alpha)\)
```

    * Sampling phase *
    2. Guess $y \in \mathcal{N}$
3. Guess whether $\left|C_{y}^{W}\right|<t$ :
4. If $\left|C_{y}^{W}\right|<t$ : Guess $\alpha_{y} \stackrel{\text { def }}{=}\left|C_{y}^{W}\right|$ exactly, i.e. $\alpha_{y} \in\{0,1, \ldots, t-1\}$
5. Otherwise: Guess $\nu \cdot n_{M} \leqslant \alpha_{y} \leqslant n_{M}$ such that $\left|C_{y}^{W}\right| \leqslant \alpha_{y} \leqslant\left(1+\frac{\varepsilon}{6}\right)\left|C_{y}^{W}\right|$
6. If $\alpha_{y}=0:\left(U^{\left(y, \alpha_{y}\right)}, V^{\left(y, \alpha_{y}\right)}\right)=$ EstimateBestResponse $_{A, k}(\widetilde{C}, \alpha)$
7. Else
8. Sample u.a.r. $\min \left\{t, \alpha_{y}\right\}$ columns from $M$; let $\widetilde{C}_{y}$ be the resulting multiset ${ }^{4}$
9. Sample u.a.r. one column from $M$; call it $\widetilde{R}_{y}$
10. $\left(U^{\left(y, \alpha_{y}\right)}, V^{\left(y, \alpha_{y}\right)}\right)=\operatorname{Sample}_{A, k, \varepsilon, t}\left(M, \mathcal{N} \backslash\{y\}, \widetilde{R} \cup\left\{\widetilde{R}_{y}\right\}, \widetilde{C} \cup\left\{\widetilde{C}_{y}\right\}, \alpha \cup\left\{\alpha_{y}\right\}\right)$
11. Let $\left(U^{(1)}, V^{(1)}\right)$ be the pair minimizing $\left\|A-U^{\left(y, \alpha_{y}\right)} \cdot V^{\left(y, \alpha_{y}\right)}\right\|_{0}$ over all guesses $y$ and $\alpha_{y}$
    * Pruning phase *
12. Let $M^{\prime}$ be matrix $M$ after the deleting $n_{M} / 2$ closest columns to $\widetilde{R}$,
i.e., the $n_{M} / 2$ columns $M_{:, j}$ with smallest values $\min _{y \in\{0,1\}^{k} \backslash \mathcal{N}}\left\|M_{:, j}-\widetilde{R}_{y}\right\|_{0}$
13. $\left(U^{(2)}, V^{(2)}\right)=\operatorname{Sample}_{A, k, \varepsilon, t}\left(M^{\prime}, \mathcal{N}, \widetilde{R}, \widetilde{C}, \alpha\right)$
    * Decision *
14. Return $\left(U^{(\ell)}, V^{(\ell)}\right)$ with the minimal value $\left\|A-U^{(\ell)} \cdot V^{(\ell)}\right\|_{0}$ over $\ell \in\{1,2\}$.
[^3]```
Algorithm 3 Estimating Best Response
EstimateBestResponse \({ }_{A, k}(\widetilde{C}, \alpha)\)
1. (Compute \(\widetilde{U})\) Compute a matrix \(\widetilde{U}=\widetilde{U}(\widetilde{C}, \alpha)\), where row \(\widetilde{U}_{i, \text { : }}\) is any vector \(x\) minimizing the estimated
\(\operatorname{cost} \widetilde{E}_{i, x}\). Note that each row \(\widetilde{U}_{i,:} \in\{0,1\}^{k}\) can be optimized independently.
2. (Compute \(\widetilde{V}\) ) Compute \(\widetilde{V}\) as a best response to \(\widetilde{U}\).
3. Return \((\widetilde{U}, \widetilde{V})\)
```


### 2.3.3 Correctness of Algorithm Sample

With notation as above, we now prove correctness of algorithm Sample.
Theorem 2.11. Algorithm Sample ${ }_{A, k, \varepsilon, t}$ generates a recursion tree which with probability at least $\left(\frac{\varepsilon}{2 t}\right)^{2^{O(k)} \cdot t}$ has a leaf calling EstimateBestResponse ${ }_{A, k}(\widetilde{C}, \alpha)$ such that
(i) $\alpha$ is a sequence of $\frac{\varepsilon}{6}$-approximate cluster sizes (w.r.t. the fixed matrix $W$ ), and
(ii) $\widetilde{C}$ is drawn according to distribution $\mathcal{D}_{W, t}$.

The rest of this section is devoted to proving Theorem 2.11. Similarly as in the algorithm, we define parameters

$$
\gamma \stackrel{\text { def }}{=} \varepsilon / 2^{k+4} \quad \text { and } \quad \nu_{i} \stackrel{\text { def }}{=} \gamma^{2^{k}+2-i} .
$$

Sort $\{0,1\}^{k}=\left\{y_{1}, \ldots, y_{2^{k}}\right\}$ such that $\left|C_{y_{1}}^{W}\right| \leqslant \ldots \leqslant\left|C_{y_{2 k}}^{W}\right|$. We construct the leaf guaranteed by the theorem inductively. In each depth $d=0,1, \ldots$ we focus on one recursive call, see Algorithm 2,

$$
\text { Sample }_{A, k, \varepsilon, t}\left(M^{(d)}, \mathcal{N}^{(d)}, \widetilde{R}^{(d)}, \widetilde{C}^{(d)}, \alpha^{(d)}\right)
$$

We consider the partitioning $P^{(d)} \stackrel{\text { def }}{=}\left\{P_{y}^{(d)}\right\}_{y \in\{0,1\}^{k}}$ induced by the partitioning $C^{W}$ on $M^{(d)}$, i.e., $P_{y}^{(d)}$ is the set $C_{y}^{W}$ restricted to the columns of $A$ that appear in the submatrix $M^{(d)}$. We claim that we can find a root-to-leaf path such that the following inductive invariants hold with probability at least $\left(\nu_{0} / t\right)^{\left(2^{k}-\left|\mathcal{N}^{(d)}\right|\right)(t+1)}$ :

I1. $P_{y}^{(d)}=C_{y}^{W}$ for all $y \in \mathcal{N}^{(d)}$, i.e., no column of an unsampled cluster has been removed,
I2. $\mathcal{N}^{(d)}=\left\{y_{1}, \ldots, y_{\mid \mathcal{N}^{(d) \mid}}\right\}$, i.e., the remaining clusters are the $\left|\mathcal{N}^{(d)}\right|$ smallest clusters,
I3. For any $y \in\{0,1\}^{k} \backslash \mathcal{N}^{(d)}$ the value $\alpha_{y}^{(d)}$ is an $\frac{\varepsilon}{6}$-approximate cluster size, i.e., if $\left|C_{y}^{W}\right|<t$ we have $\alpha_{y}^{(d)}=\left|C_{y}^{W}\right|$, and otherwise $\left|C_{y}^{W}\right| \leqslant \alpha_{y}^{(d)} \leqslant\left(1+\frac{\varepsilon}{6}\right)\left|C_{y}^{W}\right|$,

I4. For any $y \in\{0,1\}^{k} \backslash \mathcal{N}^{(d)}$ the multiset $\widetilde{C}_{y}^{(d)}$ is sampled according to distribution $\mathcal{D}_{W, t}$, i.e., if $\left|C_{y}^{W}\right|<t$ then $\widetilde{C}_{y}^{(d)}=C_{y}^{W}$ and otherwise $\widetilde{C}_{y}^{(d)}$ consists of $t$ uniformly random samples from $C_{y}^{W}$ with replacement, and

I5. For any $y \in\{0,1\}^{k} \backslash \mathcal{N}^{(d)}$ the vector $\widetilde{R}_{y}^{(d)}$ satisfies $\left\|\widetilde{R}_{y}^{(d)}-U \cdot y\right\|_{0} \leqslant 2\|A-U \cdot W\|_{0} /\left|C_{y}^{W}\right|$.
For shorthand, we let $n^{(d)} \stackrel{\text { def }}{=} n_{M^{(d)}}$ and $\nu^{(d)} \stackrel{\text { def }}{=} \nu_{\left|\mathcal{N}^{(d)}\right|}$.

Base Case: Note that the recursion may stop in Step 1 with $\mathcal{N}^{(d)}=\emptyset$ or $n^{(d)}=0$, or in Step 6 with $\alpha_{y}^{(d)}=0$ for some guessed $y \in \mathcal{N}$. Since we only want to show existence of a leaf of the recursion tree, in the latter case we can assume that we guessed $y=y_{\left|\mathcal{N}^{(d)}\right|}$ and $\alpha_{y}^{(d)}=\left|C_{y}^{W}\right|$, and thus we have $\left|C_{y}^{W}\right|=0$. Hence, in all three cases we have $\left|C_{y}^{W}\right|=0$ for all $y \in \mathcal{N}^{(d)}$, by invariant I2 and sortedness of $y_{1}, \ldots, y_{2^{k}}$. Since we initialize $\widetilde{C}_{y}^{(0)}=\emptyset$ and $\alpha_{y}^{(0)}=0$, we are done for all $y \in \mathcal{N}^{(d)}$. By invariants I3 and I4, we are also done for all $y \in\{0,1\}^{k} \backslash \mathcal{N}^{(d)}$. The total success probability is at least

$$
\left(\frac{\nu_{0}}{t}\right)^{2^{k}(t+1)}=\left(\frac{\varepsilon}{2^{k+4} t}\right)^{2^{k}\left(2^{k}+2\right)(t+1)}=\left(\frac{\varepsilon}{2 t}\right)^{2^{O(k)} \cdot t} .
$$

The proof of the inductive step proceeds by case distinction.

Case 1 (Sampling): Suppose $\left|P_{y}^{(d)}\right| \geqslant \nu^{(d)} \cdot n^{(d)}$ for some $y \in \mathcal{N}^{(d)}$. Since $P_{y}^{(d)}=C_{y}^{W}$ (by invariant I1) and sortedness, we have $\left|C_{y}^{W}\right| \geqslant \nu^{(d)} n^{(d)}$ for $y \stackrel{\text { def }}{=} y_{\left|\mathcal{N}^{(d)}\right|}$. We may assume that we guess $y=y_{\left|\mathcal{N}^{(d) \mid}\right|}$ in Step 2, since we only want to prove existence of a leaf of the recursion tree. Note that there is a number $\nu^{(d)} n^{(d)} \leqslant \alpha_{y} \leqslant n^{(d)}$ with $\left|C_{y}^{W}\right| \leqslant \alpha_{y} \leqslant\left(1+\frac{\varepsilon}{6}\right)\left|C_{y}^{W}\right|$ (in particular $\alpha_{y}=\left|C_{y}^{W}\right|$ would work), so we can guess such a number in Step 5. Together with Steps 3 and 4, we can assume that $\alpha^{(d+1)}$ satisfies invariant I3.

In Step 8 we sample a multiset $\widetilde{C}_{y}$ of $\min \left\{t, \alpha_{y}\right\}$ columns from $M$. If $\left|C_{y}^{W}\right| \geqslant t$, we condition on the event that all these columns lie in $C_{y}^{W}$. Then $\widetilde{C}_{y}$ forms a uniform sample from $C_{y}^{W}$ of size $t$. Since $\left|C_{y}^{W}\right| \geqslant \nu^{(d)} n^{(d)}$, this event has probability at least $\left(\nu^{(d)}\right)^{t}$. Otherwise, if $\left|C_{y}^{W}\right|=\alpha_{y}<t$, we condition on the event that all $\alpha_{y}$ samples lie in $C_{y}^{W}$ and are distinct. Then $\widetilde{C}_{y}=C_{y}^{W}$. The probability of this event is at least

$$
\left(\frac{1}{n^{(d)}}\right)^{\alpha_{y}} \geqslant\left(\frac{\nu^{(d)}}{\alpha_{y}}\right)^{\alpha_{y}} \geqslant\left(\frac{\nu^{(d)}}{t}\right)^{t}
$$

In total, $\widetilde{C}^{(d+1)}$ satisfies invariant I4 with probability at least $\left(\nu^{(d)} / t\right)^{t}$.
In Step 9 we sample one column $\widetilde{R}_{y}$ uniformly at random from $M$. With probability at least $\nu^{(d)}, \widetilde{R}_{y}$ belongs to $C_{y}^{W}$, and conditioned on this event $\mathcal{E}_{y}$ we have

$$
\mathbb{E}_{\widetilde{R}_{y}}\left[\left\|\widetilde{R}_{y}-U \cdot y\right\|_{0} \mid \mathcal{E}_{y}\right]=\frac{1}{\left|C_{y}^{W}\right|} \sum_{j \in C_{y}^{W}}\left\|A_{:, j}-U \cdot y\right\|_{0} \leqslant \frac{\|A-U \cdot W\|_{0}}{\left|C_{y}^{W}\right|}
$$

By Markov's inequality, with probability at least $\nu^{(d)} / 2$ we have $\left\|\widetilde{R}_{y}-U \cdot y\right\|_{0} \leqslant 2\|A-U \cdot W\|_{0} /\left|C_{y}^{W}\right|$, and thus invariant $I 5$ holds for $\widetilde{R}^{(d+1)}$.

Finally, since we did not change $M^{(d)}$, invariant I1 is maintained. We conditioned on events that hold with combined probability at least $\left(\nu^{(d)} / t\right)^{t} \cdot \nu^{(d)} / 2 \geqslant\left(\nu_{0} / t\right)^{t+1}$. Since we decrement $\left|\mathcal{N}^{(d)}\right|$ by removing $y=y_{\left|\mathcal{N}^{(d)}\right|}$ from $\mathcal{N}^{(d)}$, we maintain invariant I2, and we obtain total probability at least $\left(\nu_{0} / t\right)^{\left(2^{k}-\left|\mathcal{N}^{(d+1)}\right|\right)(t+1)}$.

Case 2 (Pruning): Suppose $\left|P_{y}^{(d)}\right|<\nu^{(d)} \cdot n^{(d)}$ for every $y \in \mathcal{N}^{(d)}$. (Note that cases 1 and 2 are complete.) In this case, we remove the $n^{(d)} / 2$ columns of $M^{(d)}$ that are closest to $\widetilde{R}^{(d)}$, resulting in a matrix $M^{(d+1)}$, and then start a recursive call on $M^{(d+1)}$. Since we do not change $\mathcal{N}^{(d)}, \widetilde{R}^{(d)}, \widetilde{C}^{(d)}$, and $\alpha^{(d)}$, invariants I2-I5 are maintained.

Invariant I1 is much more difficult to verify, as we need to check that the $n^{(d)} / 2$ deleted columns do not contain any column from an unsampled cluster. We first show that some column of a cluster we already sampled from survives to depth $d+1$ and has small distance to $\widetilde{R}^{(d)}$ (see Claim 2.12). Then we show that every column of a cluster that we did not yet sample from has large distance to $\widetilde{R}^{(d)}$ (see Claim 2.14). Since we delete the $n^{(d)} / 2$ closest columns to $\widetilde{R}^{(d)}$, it follows that every column of a cluster that we did not yet sample from survives.

Claim 2.12. There exists $x \in\{0,1\}^{k} \backslash \mathcal{N}^{(d)}$ and column $j \in P_{x}^{(d+1)}$ with

$$
\left\|A_{:, j}-\widetilde{R}_{x}^{(d)}\right\|_{0} \leqslant 2^{k+4}\|A-U \cdot W\|_{0} / n^{(d)}
$$

Proof. By Case 2, we have $\left|P_{y}^{(d)}\right|<\nu^{(d)} \cdot n^{(d)}$ for every $y \in \mathcal{N}^{(d)}$, and since $\nu^{(d)} \leqslant \nu_{2^{k}} \leqslant 2^{-k-2}$ it follows that

$$
\sum_{y \in \mathcal{N}^{(d)}}\left|P_{y}^{(d)}\right|<2^{k} \nu^{(d)} n^{(d)} \leqslant n^{(d)} / 4
$$

Since $\left|P_{y}^{(d)}\right| \geqslant\left|P_{y}^{(d+1)}\right|$ and $\sum_{y \in\{0,1\}^{k}}\left|P_{y}^{(d+1)}\right|=n^{(d)} / 2$, we obtain $\sum_{y \in\{0,1\}^{k} \backslash \mathcal{N}^{(d)}}\left|P_{y}^{(d+1)}\right| \geqslant n^{(d)} / 4$. Hence, there is $x \in\{0,1\}^{k} \backslash \mathcal{N}^{(d)}$ such that

$$
\begin{equation*}
\left|P_{x}^{(d+1)}\right| \geqslant 2^{-k-2} n^{(d)} \tag{2.14}
\end{equation*}
$$

By the minimum-arithmetic-mean inequality, there is $j \in P_{x}^{(d+1)}$ such that

$$
\begin{aligned}
\left\|A_{:, j}-\widetilde{R}_{x}^{(d)}\right\|_{0} & \leqslant \frac{1}{\left|P_{x}^{(d+1)}\right|} \cdot \sum_{j^{\prime} \in P_{x}^{(d+1)}}\left\|A_{:, j^{\prime}}-\widetilde{R}_{x}^{(d)}\right\|_{0} \\
& \leqslant\left\|\widetilde{R}_{x}^{(d)}-U \cdot x\right\|_{0}+\frac{1}{\left|P_{x}^{(d+1)}\right|} \cdot \sum_{j^{\prime} \in P_{x}^{(d+1)}}\left\|A_{:, j^{\prime}}-U \cdot x\right\|_{0}
\end{aligned}
$$

where the last step uses the triangle inequality. For the first summand we use invariant I5, and for the second we use that $P_{x}^{(d+1)}$ is by definition part of an induced partitioning of $C^{W}$ on a smaller matrix, and thus the summation is bounded by $\|A-U W\|_{0}$. This yields

$$
\left\|A_{:, j}-\widetilde{R}_{x}^{(d)}\right\|_{0} \leqslant\left(\frac{2}{\left|C_{x}^{W}\right|}+\frac{1}{\left|P_{x}^{(d+1)}\right|}\right) \cdot\|A-U \cdot W\|_{0}
$$

By $P_{x}^{(d+1)} \subseteq C_{x}^{W}$ and by (2.14), we obtain the claimed bound.
Claim 2.13. For any $y \in\{0,1\}^{k} \backslash \mathcal{N}^{(d)}$ we have $\left|C_{y}^{W}\right| \geqslant \nu^{(d)} n^{(d)} / \gamma$.
Proof. Since $y \notin \mathcal{N}$, we sampled from this cluster in some depth $d^{\prime}<d$. In the call corresponding to $d^{\prime}$, we had $\mathcal{N}^{\left(d^{\prime}\right)} \supseteq \mathcal{N}^{(d)} \cup\{y\}$ and thus $\left|\mathcal{N}^{\left(d^{\prime}\right)}\right| \geqslant\left|\mathcal{N}^{(d)}\right|+1$, and we had $n^{\left(d^{\prime}\right)} \geqslant n^{(d)}$. Since we sampled from $C_{y}^{W}$ in depth $d^{\prime}$, Case 1 was applicable, and thus

$$
\begin{aligned}
\left|C_{y}^{W}\right| & \stackrel{(\mathrm{I} 1)}{=}\left|P_{y}^{\left(d^{\prime}\right)}\right| \geqslant \nu^{\left(d^{\prime}\right)} \cdot n^{\left(d^{\prime}\right)}=\nu_{\left|\mathcal{N}^{\left(d^{\prime}\right)}\right|} \cdot n^{\left(d^{\prime}\right)} \\
& \geqslant \nu_{\left|\mathcal{N}^{d}\right|+1} \cdot n^{(d)}=\nu_{\left|\mathcal{N}^{d}\right|} \cdot n^{(d)} / \gamma=\nu^{(d)} \cdot n^{(d)} / \gamma
\end{aligned}
$$

Claim 2.14. For any $y \in\{0,1\}^{k} \backslash \mathcal{N}^{(d)}$, $z \in \mathcal{N}^{(d)}$, and $j \in P_{z}^{(d)}$ we have

$$
\left\|A_{:, j}-\widetilde{R}_{y}^{(d)}\right\|_{0}>2^{k+4}\|A-U \cdot W\|_{0} / n^{(d)}
$$

Proof. By triangle inequality, we have $\|U \cdot y-U \cdot z\|_{0} \leqslant\left\|A_{:, j}-U \cdot y\right\|_{0}+\left\|A_{:, j}-U \cdot z\right\|_{0}$. Since $j \in P_{z}^{(d)}=C_{z}^{W}$ and by property (ii) of ( $U, V, \varepsilon$ )-clustered (see Lemma 2.10), the first summand is at least as large as the second, and we obtain

$$
\|U \cdot y-U \cdot z\|_{0} \leqslant 2\left\|A_{:, j}-U \cdot y\right\|_{0}
$$

We use this and the triangle inequality to obtain

$$
\begin{aligned}
\left\|A_{:, j}-\widetilde{R}_{y}^{(d)}\right\|_{0} & \geqslant\left\|A_{:, j}-U \cdot y\right\|_{0}-\left\|\widetilde{R}_{y}^{(d)}-U \cdot y\right\|_{0} \\
& \geqslant \frac{1}{2}\|U \cdot y-U \cdot z\|_{0}-\left\|\widetilde{R}_{y}^{(d)}-U \cdot y\right\|_{0}
\end{aligned}
$$

For the first summand we use property (i) of $(U, V, \varepsilon)$-clustered (see Lemma 2.10), for the second we use invariant I5. This yields

$$
\left\|A_{:, j}-\widetilde{R}_{y}^{(d)}\right\|_{0}>\frac{\varepsilon}{2^{k+1}} \cdot \frac{\|A-U \cdot V\|_{0}}{\left|C_{z}^{W}\right|}-\frac{2\|A-U \cdot W\|_{0}}{\left|C_{y}^{W}\right|}
$$

Since $y \in\{0,1\}^{k} \backslash \mathcal{N}^{(d)}$, Claim 2.13 yields $\left|C_{y}^{W}\right| \geqslant \nu^{(d)} n^{(d)} / \gamma$. Since $z \in \mathcal{N}^{(d)}$, by invariant I1, and since we are in Case 2, we have $\left|C_{z}^{W}\right|=\left|P_{z}^{(d)}\right|<\nu^{(d)} \cdot n^{(d)}$. Moreover, by the properties of $(U, V, \varepsilon)$-clustered (see Lemma 2.10), we have $\|A-U W\|_{0} \leqslant(1+\varepsilon)\|A-U \cdot V\|_{0}$ and thus $\|A-U \cdot V\|_{0} \geqslant \frac{1}{2}\|A-U \cdot W\|_{0}$. Together, this yields

$$
\begin{aligned}
\left\|A_{:, j}-\widetilde{R}_{y}^{(d)}\right\|_{0} & >\left(\frac{\varepsilon}{2^{k+2}}-2 \gamma\right) \cdot \frac{\|A-U \cdot W\|_{0}}{\nu^{(d)} n^{(d)}} \\
& =\frac{\varepsilon}{2^{k+3} \nu^{(d)}} \cdot \frac{\|A-U \cdot W\|_{0}}{n^{(d)}} \\
& \geqslant 2^{k+4} \cdot \frac{\|A-U \cdot W\|_{0}}{n^{(d)}}
\end{aligned}
$$

since $\gamma=\varepsilon / 2^{k+4}$ and $\nu^{(d)} \leqslant \nu_{2^{k}}=\gamma^{2} \leqslant \varepsilon / 2^{2 k+7}$.
Together, Claims 2.12 and 2.14 prove that no column $j \in P_{y}^{(d)}$ with $y \in \mathcal{N}^{(d)}$ is removed. Indeed, we remove the columns with smallest distance to $\widetilde{R}^{(d)}$, some column in distance $2^{k+4}\|A-U W\|_{0} / n^{(d)}$ survives, and any column $j \in P_{y}^{(d)}$ with $y \in \mathcal{N}$ has larger distance to $\widetilde{R}^{(d)}$. It follows that invariant I1 is maintained, completing our proof of correctness.

### 2.3.4 Running Time Analysis of Algorithm Sample

We now analyze the running time of Algorithm 3.
Lemma 2.15. Algorithm EstimateBestResponse runs in time $2^{O(k)} \mathrm{mn}$.
Proof. Note that if $\widetilde{C}$ is drawn according to distribution $\mathcal{D}_{W, t}$, then its total size $\sum_{y \in\{0,1\}^{k}}\left|\widetilde{C}_{y}\right|$ is at most $n$. Hence, we can ignore all calls violating this inequality. We can thus evaluate the estimated cost $\widetilde{E}_{i, x}$ in time $2^{O(k)} n$. Optimizing over all $x \in\{0,1\}^{k}$ costs another factor $2^{k}$, and iterating over all rows $i$ adds a factor $m$. Thus, Step 1 runs in time $2^{O(k)} m n$. Further, Step 2 finds a best response matrix, which can be computed in the same running time.

We proceed by analyzing the time complexity of Algorithm 2.
Lemma 2.16. For any $t=\operatorname{poly}\left(2^{k} / \varepsilon\right)$, Algorithm Sample $_{A, k, \varepsilon, t}$ runs in time $(2 / \varepsilon)^{2^{O(k)}} \cdot m n^{1+o(1)}$, where $o(1)$ hides a factor $(\log \log n)^{1.1} / \log n$.
Proof. Consider any recursive call $\operatorname{Sample}_{A, k, \varepsilon, t}(M, \mathcal{N}, \widetilde{R}, \widetilde{C}, \alpha)$. We express its running time as $T(a, b)$ where $a \stackrel{\text { def }}{=}|\mathcal{N}|$ and $b \stackrel{\text { def }}{=} \log \left(n_{M}\right)$. For notational convenience, we let $\log (0)=:-1$ and assume that $n_{M}$ is a power of 2 .

If we make a call to algorithm EstimateBestResponse then this takes time $2^{O(k)} \cdot m n$ by the preceding lemma. Note that here we indeed have the size $n$ of the original matrix and not the size $n_{M}$ of the current submatrix, since we need to determine the cost with respect to the original matrix.

In the sampling phase, in Step 2 we guess $y$, with $|\mathcal{N}| \leqslant 2^{k}$ possibilities. Moreover, in Steps 3-5 we guess either $\alpha_{y} \in\{0,1, \ldots, t-1\}$ or $\nu_{|\mathcal{N}|} \cdot n_{M} \leqslant \alpha_{y} \leqslant n_{M}$ such that $\left|C_{y}^{W}\right| \leqslant \alpha_{y} \leqslant\left(1+\frac{\varepsilon}{6}\right)\left|C_{y}^{W}\right|$. Note that there are $O\left(\log \left(1 / \nu_{|\mathcal{N}|}\right) / \log (1+\varepsilon / 6)\right)=\operatorname{poly}\left(2^{k} / \varepsilon\right)$ possibilities for the latter, and thus poly $\left(2^{k} / \varepsilon\right)$ possibilities in total. For each such guess we make one recursive call with a decremented $a$ and we evaluate the cost of the returned solution in time $2^{O(k)} \cdot m n$.

In the pruning phase, we delete the $n_{M} / 2$ columns that are closest to $\widetilde{R}$, which can be performed in time $2^{O(k)} \cdot m \cdot n_{M}$ (using median-finding in linear time). We then make one recursive call with a decremented $b$.

Together, we obtain the recursion

$$
T(a, b) \leqslant \operatorname{poly}\left(2^{k} / \varepsilon\right) m n+\operatorname{poly}\left(2^{k} / \varepsilon\right) \cdot T(a-1, b)+T(a, b-1)
$$

with base cases $T(0, b)=T(a,-1)=2^{O(k)} m n$. The goal is to upper bound $T\left(2^{k}, \log n\right)$.
Let $Y=\operatorname{poly}\left(2^{k} / \varepsilon\right)$ and $X=Y \cdot m n$ such that

$$
T(a, b) \leqslant X+Y \cdot T(a-1, b)+T(a, b-1)
$$

and $T(0, b), T(a,-1) \leqslant X$. We prove by induction that $T(a, b) \leqslant X \cdot(2 Y(b+2))^{a}$. This works in the base cases where $a=0$ or $b=-1$. Inductively, for $a>0$ and $b \geqslant 0$ we bound ${ }^{5}$

$$
\begin{align*}
T(a, b) & \leqslant X+Y \cdot X \cdot(2 Y(b+2))^{a-1}+X \cdot(2 Y(b+1))^{a} \\
& =X \cdot(2 Y(b+2))^{a} \cdot\left(\frac{1}{(2 Y(b+2))^{a}}+\frac{1}{2(b+2)}+\left(\frac{b+1}{b+2}\right)^{a}\right) \\
& \leqslant X \cdot(2 Y(b+2))^{a} \cdot\left(\frac{1}{2(b+2)}+\frac{1}{2(b+2)}+\frac{b+1}{b+2}\right) \\
& =X \cdot(2 Y(b+2))^{a} . \tag{2.15}
\end{align*}
$$

Let $C$ be a constant to be determined soon. Using (2.15), the total running time is bounded by

$$
T\left(2^{k}, \log n\right) \leqslant X \cdot(2 Y(\log (n)+2))^{2^{k}} \leqslant(2 / \varepsilon)^{2^{(C+1) \cdot k}} \cdot m n \cdot \log ^{2^{k}} n \leqslant(2 / \varepsilon)^{2^{(C+1) \cdot k}} \cdot m n^{1+o(1)},
$$

where the last inequality follows by noting that $\log ^{2^{k}} n>(2 / \varepsilon)^{2^{(C+1) \cdot k}}$ iff $k<\log \left(\frac{\log \log n}{\log (2 / \varepsilon)}\right)^{1 / C}$ and in this case

$$
(\log n)^{2^{k}} \leqslant(\log n)^{\left(\frac{\log \log n}{\log (2 / \varepsilon)}\right)^{1 / C}} \leqslant n^{o(1)},
$$

where $o(1)$ hides a factor $(\log \log n)^{1+1 / C} / \log n$. The statement follows for any $C \geqslant 10$.

[^4]
### 2.3.5 The Complete PTAS: Correctness and Efficiency

Finally, we use Algorithm Sample to obtain an efficient PTAS for the Generalized Binary $\ell_{0}$-Rank- $k$ problem. Given $A, k$ and $\varepsilon$, we call Sample ${ }_{A, k, \varepsilon / 4, t}$ with

$$
t=t(k, \varepsilon / 4) \stackrel{\text { def }}{=} 2^{4 k+16} / \varepsilon^{2}
$$

(This means that we replace all occurrences of $\varepsilon$ by $\varepsilon / 4$, in particular we also assume that $W$ is $\left(U, V, \frac{\varepsilon}{4}\right)$ clusterable.) By Theorem 2.11, with probability at least

$$
\left(\frac{\varepsilon}{2 t}\right)^{2^{O(k)} \cdot t}=\left(\frac{\varepsilon}{2}\right)^{2^{O(k)} / \varepsilon^{2}}
$$

at least one leaf of the recursion tree calls EstimateBestResponse ${ }_{A, k}(\widetilde{C}, \alpha)$ with proper $\widetilde{C}$ and $\alpha$ such that the Sampling Theorem 2.3 is applicable. By choice of $t=t\left(k, \frac{\varepsilon}{4}\right)$, this yields

$$
\mathbb{E}\left[\|A-\widetilde{U}(\widetilde{C}, \alpha) \cdot W\|_{0}\right] \leqslant\left(1+\frac{\varepsilon}{4}\right) \mathrm{OPT}_{k}^{W} \leqslant\left(1+\frac{\varepsilon}{4}\right)^{2} \mathrm{OPT}_{k}
$$

where we used that $W$ is $\left(U, V, \frac{\varepsilon}{4}\right)$-clusterable in the second step (see Lemma 2.10). The Algorithm EstimateBestResponse computes a matrix $\widetilde{U}=\widetilde{U}(\widetilde{C}, \alpha)$ and its best response matrix $\widetilde{V}$. This yields

$$
\mathbb{E}\left[\|A-\widetilde{U} \cdot \tilde{V}\|_{0}\right] \leqslant \mathbb{E}\left[\|A-\widetilde{U} \cdot W\|_{0}\right] \leqslant\left(1+\frac{\varepsilon}{4}\right)^{2} \mathrm{OPT}_{k}
$$

By Markov's inequality, with probability at least $1-1 /(1+\varepsilon / 4) \geqslant \frac{\varepsilon}{5}$ we have

$$
\|A-\widetilde{U} \cdot \widetilde{V}\|_{0} \leqslant\left(1+\frac{\varepsilon}{4}\right) \cdot \mathbb{E}\left[\|A-\widetilde{U} \cdot \widetilde{V}\|_{0}\right] \leqslant\left(1+\frac{\varepsilon}{4}\right)^{3} \mathrm{OPT}_{k} \leqslant(1+\varepsilon) \mathrm{OPT}_{k}
$$

Hence, with probability at least $p=(\varepsilon / 2)^{2^{O(k)} / \varepsilon^{2}}$ at least one solution $\widetilde{U}, \widetilde{V}$ generated by our algorithm is a $(1+\varepsilon)$-approximation. Since we return the best of the generated solutions, we obtain a PTAS, but its success probability $p$ is very low.

The success probability can be boosted to a constant by running $O(1 / p)=(2 / \varepsilon)^{2^{O(k)} / \varepsilon^{2}}$ independent trials of Algorithm Sample. By Lemma 2.16, each call runs in time $(2 / \varepsilon)^{2^{O(k)}} \cdot m n^{1+o(1)}$, where $o(1)$ hides a factor $(\log \log n)^{1.1} / \log n$, yielding a total running time of $(2 / \varepsilon)^{2^{O(k)} / \varepsilon^{2}} \cdot m n^{1+o(1)}$. This finishes the proof of Theorem 1.1. The success probability can be further amplified to $1-\delta$ for any $\delta>0$, by running $O(\log (1 / \delta))$ independent trials of the preceding algorithm.

### 2.4 Faster Binary $\ell_{0}$-Rank-1

In this section, we consider the Binary $\ell_{0}$-Rank- 1 problem with standard inner product. Given a binary matrix $A \in\{0,1\}^{m \times n}$ and an error $\varepsilon \in(0,1 / 2)$, our goal is to find binary vectors $\widetilde{u} \in\{0,1\}^{m}, \widetilde{v} \in\{0,1\}^{n}$ such that $\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0} \leqslant(1+\varepsilon)$ OPT, where the optimal value is defined by

$$
\mathrm{OPT}=\min _{u \in\{0,1\}^{m}, v \in\{0,1\}^{n}}\left\|A-u \cdot v^{T}\right\|_{0}
$$

We now give a faster PTAS for the Binary $\ell_{0}$-Rank-1 problem, which improves upon Theorem 1.1.
Theorem 1.3 (from page 5). (PTAS for the Binary $\ell_{0}-R a n k-1$ problem with standard inner product) For any $\varepsilon \in(0,1 / 2)$ there is an algorithm that runs in time $(1 / \varepsilon)^{O\left(1 / \varepsilon^{2}\right)} \cdot\left(\|A\|_{0}+m+n\right) \cdot \log ^{3}(m n)$, and outputs vectors $\widetilde{u} \in\{0,1\}^{m}, \widetilde{v} \in\{0,1\}^{n}$ such that w.h.p. ${ }^{6}\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0} \leqslant(1+\varepsilon) \min _{u \in\{0,1\}^{m}, v \in\{0,1\}^{n}}\left\|A-u \cdot v^{T}\right\|_{0}$.

The remainder of this section is devoted to proving Theorem 1.3.

[^5]Structural Properties Let $R_{1}=\left\{i \in[m]: u_{i}=1\right\}$ and $C_{1}=\left\{j \in[n]: v_{j}=1\right\}$. Then, estimated cost simplifies as follows: for any row $i$ and binary scalar $x \in\{0,1\}$, we set

$$
E_{i, x}^{\prime} \stackrel{\text { def }}{=}\left|\left\{j \in \widetilde{C}_{1}: A_{i, j} \neq x\right\}\right| .
$$

Lemma 2.17. For any row $i$ and $x, x^{\prime} \in\{0,1\}$ we have $\widetilde{E}_{i, x}<\widetilde{E}_{i, x^{\prime}}$ if and only if $E_{i, x}^{\prime}<E_{i, x^{\prime}}^{\prime}$.

Proof. By definition of $\widetilde{E}_{i, x}$ and by $k=1$, we have $\widetilde{E}_{i, x}<\widetilde{E}_{i, x^{\prime}}$ if and only if

$$
\sum_{y \in\{0,1\}} \frac{\left|C_{y}\right|}{\left|\widetilde{C}_{y}\right|} \cdot\left|\left\{j \in \widetilde{C}_{y}: A_{i, j} \neq\langle x, y\rangle\right\}\right|<\sum_{y \in\{0,1\}} \frac{\left|C_{y}\right|}{\left|\widetilde{C}_{y}\right|} \cdot\left|\left\{j \in \widetilde{C}_{y}: A_{i, j} \neq\left\langle x^{\prime}, y\right\rangle\right\}\right| .
$$

For $y=0$ we have $\langle x, y\rangle=0=\left\langle x^{\prime}, y\right\rangle$, and thus the contribution of the corresponding summand to both sides is equal. Hence, we have, equivalently,

$$
\frac{\left|C_{1}\right|}{\left|\widetilde{C}_{1}\right|} \cdot\left|\left\{j \in \widetilde{C}_{1}: A_{i, j} \neq x\right\}\right|<\frac{\left|C_{1}\right|}{\left|\widetilde{C}_{1}\right|} \cdot\left|\left\{j \in \widetilde{C}_{1}: A_{i, j} \neq x^{\prime}\right\}\right|
$$

Removing the common factor $\left|C_{1}\right| /\left|\widetilde{C}_{1}\right|$, we equivalently arrive at $E_{i, x}^{\prime}<E_{i, x^{\prime}}^{\prime}$. This also holds in the special case $\left|\widetilde{C}_{1}\right|=0$, since then $\widetilde{E}_{i, x}=\widetilde{E}_{i, x^{\prime}}$ and $E_{i, x}^{\prime}=E_{i, x^{\prime}}^{\prime}=0$.

Sampling Theorem Recall that we choose the matrix $\widetilde{U}$ by picking for each row $i$ a vector $\widetilde{U}_{i,:}=x$ minimizing $\widetilde{E}_{i, x}$. The above lemma shows that we can equivalently minimize $E_{i, x}^{\prime}$. Let us formulate the resulting sampling theorem.

Corollary 2.18. For any $\varepsilon \in(0,1 / 2)$ set $t \stackrel{\text { def }}{=} t(1, \varepsilon / 2)=2^{16} / \varepsilon^{2}$. If $\left|C_{1}\right|<t$ then set $\widetilde{C}_{1} \stackrel{\text { def }}{=} C_{1}$, otherwise sample $t$ elements from $C_{1}$ with replacement and let the resulting multiset be $\widetilde{C}_{1}$. Construct $\widetilde{u} \in\{0,1\}^{m}$ by picking $\widetilde{u}_{i}=x \in\{0,1\}$ minimizing $E_{i, x}^{\prime}=\left|\left\{j \in \widetilde{C}_{1}: A_{i, j} \neq x\right\}\right|$. Then we have

$$
\mathbb{E}_{\widetilde{C}_{1}}\left[\left\|A-\widetilde{u} \cdot v^{T}\right\|_{0}\right] \leqslant\left(1+\frac{\varepsilon}{2}\right) \text { OPT. }
$$

Proof. We obtain this construction of $\widetilde{u}$ as follows: Specialize the construction in Section 2.2.2 to $k=1$, replace $\widetilde{E}_{i, x}$ by $E_{i, x}^{\prime}$, which does not change the result by Lemma 2.17 , and finally remove redundant steps like the sampling of $\widetilde{C}_{0}$, since it is no longer used in the construction. The conclusion thus follows from Theorem 2.3. For the purpose of this section, we reduced the approximation ratio from $1+\varepsilon$ to $1+\varepsilon / 2$, thereby increasing $t$ by a factor 4 , see (2.13).

### 2.4.1 The Algorithm

We present now our efficient randomized PTAS for the Binary $\ell_{0}$-Rank-1 problem. The algorithm succeeds with probability at least $\left(\frac{\varepsilon}{5 t}\right)^{t+1}$.

```
Algorithm 4 Faster randomized PTAS for the Binary \(\ell_{0}\)-Rank-1 problem
Input: A matrix \(A \in\{0,1\}^{m \times n}\) and error \(\varepsilon \in(0,1)\).
Output: Vectors \(\widetilde{u} \in\{0,1\}^{m}\) and \(\widetilde{v} \in\{0,1\}^{n}\) satisfying \(\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0} \leqslant(1+\varepsilon)\) OPT.
```

1. (Basic solution) Initialize $S=\left\{\left(0_{m}, 0_{n}\right)\right\}$, where $0_{d}=(0, \ldots, 0) \in\{0,1\}^{d}$.
2. (Guess approximate column set size) Exhaustively guess $\left\lceil\left|C_{1}\right|\right\rceil_{2}$, where $\lceil r\rceil_{2}$ denotes the smallest power of 2 that is at least $r$. There are $O(\log n)$ possibilities. If we guessed $\left\lceil\left|C_{1}\right|\right\rceil_{2} \leqslant\lceil t\rceil_{2}$, then exhaustively guess $\left|C_{1}\right|$ exactly. There are $O(t)$ possibilities. In particular, if we guessed correctly then we know the number $\min \left\{t,\left|C_{1}\right|\right\}$.
3. (Guess approximate row set size) Exhaustively guess $\left\lceil\left|R_{1}\right|\right\rceil_{2}$ where $R_{1} \stackrel{\text { def }}{=}\left\{i: u_{i}=1\right\}$. There are $O(\log m)$ possibilities.
4. (Ignore sparse columns) Let $W \stackrel{\text { def }}{=}\left\{j:\left\|A_{:, j}\right\|_{0} \geqslant\left\lceil\left|R_{1}\right|\right\rceil_{2} / 4\right\}$ be the set of columns of $A$ containing at least $\left\lceil\left|R_{1}\right|\right\rceil_{2} / 4$ ones. Computing $W$ takes time $O\left(\|A\|_{0}\right)$.
5. (Sample columns) Sample a multiset $C_{1}^{\prime}$ consisting of $\min \left\{t,\left|C_{1}\right|\right\}$ columns chosen independently and uniformly at random from $W$. This takes time $O(t)$.
6. (Compute $\widetilde{u}$ ) We plug $C_{1}^{\prime}$ as $\widetilde{C}_{1}$ into the estimated $\operatorname{cost} E_{i, x}^{\prime}=\left|\left\{j \in \widetilde{C}_{1}: A_{i, j} \neq x\right\}\right|$. Compute a vector $\widetilde{u}$ by picking $\widetilde{u}_{i}=x \in\{0,1\}$ minimizing the cost $E_{i, x}^{\prime}$, for all $i$. This takes time $O(m t)$ since $\left|C_{1}^{\prime}\right| \leqslant t$.
7. (Compute $\widetilde{v}$ ) Compute $\widetilde{v}$ as a best response to $\widetilde{u}$. Add $(\widetilde{u}, \widetilde{v})$ to $S$.
8. Return the pair $(\widetilde{u}, \widetilde{v}) \in S$ minimizing $\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0}$ over all exhaustive guesses and the basic solution.

Note that the claimed success probability of Algorithm 4 is quite low. In order to prove Theorem 1.3 we will boost this probability by repeating the algorithm sufficiently often. In the remainder of this section we analyze the running time and the success probability of Algorithm 4.

### 2.4.2 Running Time

The only steps without immediate time bounds are Step 7 and Step 8.
For Step 7, we now argue that a best response $\widetilde{v}$ to $\widetilde{u}$ can be computed in time $O\left(\|A\|_{0}+m+n\right)$. Observe that the cost corresponding to column $A_{:, j}$ has a fixed cost term $\mid\left\{i: A_{i, j}=1\right.$ and $\left.\widetilde{u}_{i}=0\right\} \mid$ which is independent of the choice of $\widetilde{v}_{j} \in\{0,1\}$. Further, by enumerating all non-zero entries of $A$, we can determine for each $j$ the number $r_{j}=\mid\left\{i: A_{i, j}=1\right.$ and $\left.\widetilde{u}_{i}=1\right\} \mid$. If $r_{j}>\|\widetilde{u}\|_{0} / 2$ we set $\widetilde{v}_{j} \xlongequal{\text { def }} 1$, and $\widetilde{v}_{j} \stackrel{\text { def }}{=} 0$ otherwise. Straightforward checking shows that this yields a best response.

In Step 8 we need to calculate $\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0}$ for each $(\widetilde{u}, \widetilde{v}) \in S$, in order to pick the best pair. For each $(\widetilde{u}, \widetilde{v})$, this value can be computed in time $O\left(\|A\|_{0}+m+n\right)$ as follows. Enumerate all non-zero entries of $A$ to determine the numbers $p \stackrel{\text { def }}{=} \mid\left\{(i, j): A_{i, j}=1\right.$ and $\left.\widetilde{u}_{i} \cdot \widetilde{v}_{j}=1\right\} \mid$ and $q \stackrel{\text { def }}{=} \mid\left\{(i, j): A_{i, j}=1\right.$ and $\left.\widetilde{u}_{i} \cdot \widetilde{v}_{j}=0\right\} \mid$. Then, we can infer $\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0}=q+\|\widetilde{u}\|_{0} \cdot\|\widetilde{v}\|_{0}-p$.

Note that there are $O(t \log (m) \log (n))$ possibilities for all guesses (see Steps 2-3). Considering Steps $4-6$ as well as the running times for Steps $7-8$ shown above, we spend time $O\left(\|A\|_{0}+m+n t\right)=$ $O\left(\left(\|A\|_{0}+m+n\right) t\right)$ for each guess. Hence, we obtain the following.

Lemma 2.19. The running time of Algorithm 4 is $O\left(t^{2}\left(\|A\|_{0}+m+n\right) \log (n) \log (m)\right)$.

### 2.4.3 Correctness and Success Probability

We establish now the correctness of Algorithm 4. More precisely, we will prove a lower bound of $\left(\frac{\varepsilon}{5 t}\right)^{t+1}$ on its success probability.

Claim 2.20 (Step 4). We have $C_{1} \subseteq W$.
Proof. Let $R_{x} \stackrel{\text { def }}{=}\left\{i: u_{i}=x\right\}$ for $x \in\{0,1\}$. We split

$$
\mathrm{OPT}=\left\|A-u \cdot v^{T}\right\|_{0}=\sum_{j=1}^{n}\left\|A_{:, j}-u \cdot v_{j}\right\|_{0}
$$

where $\left\|A_{:, j}-u \cdot v_{j}\right\|_{0}$ is the cost of column $j$. If $v_{j}=0$ then this cost is $\left\|A_{:, j}\right\|_{0}$, i.e. the number of 1 's of $A_{:, j}$ in $R_{1}$ plus the number of 1's of $A_{\cdot, j}$ in $R_{0}$. If $v_{j}=1$ then the cost is $\left\|A_{:, j}-u\right\|_{0}$, i.e., the number of 0 's of $A_{:, j}$ in $R_{1}$ plus the number of 1's of $A_{:, j}$ in $R_{0}$.

Suppose $\left\|A_{:, j}\right\|_{0}<\left|R_{1}\right| / 2$. Then, the number of 1 's of $A_{:, j}$ in $R_{1}$ is less than $\left|R_{1}\right| / 2$, and the number of 0 's of $A_{:, j}$ in $R_{1}$ is larger than $\left|R_{1}\right| / 2$. Thus, $v_{j}=0$ has less cost, and $j \in C_{0}$. Hence,

$$
C_{1} \subseteq\left\{j:\left\|A_{:, j}\right\|_{0} \geqslant\left|R_{1}\right| / 2\right\} \subseteq\left\{j:\left\|A_{:, j}\right\|_{0} \geqslant\left\lceil\left|R_{1}\right|\right\rceil_{2} / 4\right\}=W
$$

Assume that all guesses in Steps 2-3 are correct, so that the multiset $C_{1}^{\prime}$ constructed in Step 5 consists of $\min \left\{t,\left|C_{1}\right|\right\}$ columns chosen independently and uniformly at random from $W$.

We define an event $\mathcal{E}$ as follows. If $\left|C_{1}\right| \geqslant t$, then event $\mathcal{E}$ asserts that $C_{1}^{\prime} \subseteq C_{1}$, i.e., all $t$ sampled columns forming $C_{1}^{\prime}$ hit $C_{1}$. If $\left|C_{1}\right|<t$, then event $\mathcal{E}$ asserts that $C_{1}^{\prime}=C_{1}$, i.e., all $\left|C_{1}\right|$ sampled columns forming $C_{1}^{\prime}$ hit $C_{1}$ and are distinct.

Claim 2.21 (Step 5). Let $\widetilde{C}_{1}$ be distributed as in Corollary 2.18. Conditioned on event $\mathcal{E}, C_{1}^{\prime}$ and $\widetilde{C}_{1}$ follow the same distribution.

Proof. If $\left|C_{1}\right|<t$ then we set $\widetilde{C}_{1}=C_{1}$, and conditioned on $\mathcal{E}$ we also have $C_{1}^{\prime}=C_{1}$. If $\left|C_{1}\right| \geqslant t$, then $\widetilde{C}_{1}$ is formed by sampling $t$ elements from $C_{1}$ with replacement. On the other hand, $C_{1}^{\prime}$ is formed by sampling $t$ elements from $W$ with replacement, which conditioned on $\mathcal{E}$ implies that all $t$ elements hit $C_{1}$. Since $C_{1} \subseteq W$ by Claim 2.20, the sampled elements are uniformly random in $C_{1}$. These processes describe the same distribution.

It follows that, conditioned on event $\mathcal{E}$, Corollary 2.18 is applicable. We use it to show:
Claim 2.22. Assuming correct guesses in Steps 2-3, the vectors $\widetilde{u}, \widetilde{v}$ computed in Steps 6-7 satisfy

$$
\operatorname{Pr}\left[\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0} \leqslant(1+\varepsilon) \mathrm{OPT}\right] \geqslant \frac{\varepsilon}{4} \cdot \operatorname{Pr}[\mathcal{E}]
$$

Proof. Since Step 6 follows the construction in Corollary 2.18, except that we replaced $\widetilde{C}_{1}$ by $C_{1}^{\prime}$, and using Claim 2.21, Corollary 2.18 yields

$$
\mathbb{E}_{C_{1}^{\prime}}\left[\left\|A-\widetilde{u} \cdot v^{T}\right\|_{0} \mid \mathcal{E}\right] \leqslant\left(1+\frac{\varepsilon}{2}\right) \mathrm{OPT}
$$

Using Markov's inequality with $\lambda \stackrel{\text { def }}{=}(1+\varepsilon) /(1+\varepsilon / 2)$ yields

$$
\operatorname{Pr}\left[\left.\left\|A-\widetilde{u} \cdot v^{T}\right\|_{0}>\lambda \cdot\left(1+\frac{\varepsilon}{2}\right) \mathrm{OPT} \right\rvert\, \mathcal{E}\right] \leqslant \frac{1}{\lambda}=1-\frac{\varepsilon / 2}{1+\varepsilon} \leqslant 1-\frac{\varepsilon}{4}
$$

or equivalently,

$$
\operatorname{Pr}\left[\left\|A-\widetilde{u} \cdot v^{T}\right\|_{0} \leqslant(1+\varepsilon) \mathrm{OPT} \mid \mathcal{E}\right] \geqslant \frac{\varepsilon}{4}
$$

Hence, we have

$$
\operatorname{Pr}\left[\left\|A-\widetilde{u} \cdot v^{T}\right\|_{0} \leqslant(1+\varepsilon) \mathrm{OPT}\right] \geqslant \operatorname{Pr}\left[\left\|A-\widetilde{u} \cdot v^{T}\right\|_{0} \leqslant(1+\varepsilon) \mathrm{OPT} \mid \mathcal{E}\right] \cdot \operatorname{Pr}[\mathcal{E}] \geqslant \frac{\varepsilon}{4} \cdot \operatorname{Pr}[\mathcal{E}] .
$$

Since in Step $7 \widetilde{v}$ is computed as a best response to $\widetilde{u}$, we have $\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0} \leqslant\left\|A-\widetilde{u} \cdot v^{T}\right\|_{0}$, and thus

$$
\operatorname{Pr}\left[\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0} \leqslant(1+\varepsilon) \mathrm{OPT}\right] \geqslant \frac{\varepsilon}{4} \cdot \operatorname{Pr}[\mathcal{E}]
$$

In order to give a lower bound on the success probability, it thus suffices to bound $\operatorname{Pr}[\mathcal{E}]$. To this end, we first bound $|W|$ in terms of $\left|C_{1}\right|$.

Claim 2.23. If $\mathrm{OPT}<\|A\|_{0} /(1+\varepsilon)$, then we have $|W|<5\left|C_{1}\right| / \varepsilon$.
Proof. As in the proof of Claim 2.20, the cost of any column $j \in C_{0}$ is $\left\|A_{:, j}\right\|_{0}$. Since any column $j \in W$ contains at least $\left\lceil\left|R_{1}\right|\right\rceil_{2} / 4 \geqslant\left|R_{1}\right| / 41$ 's, the cost of any column $j \in C_{0} \cap W$ is at least $\left|R_{1}\right| / 4$. It follows that

$$
\begin{equation*}
\left|C_{0} \cap W\right| \leqslant 5 \cdot \mathrm{OPT} /\left|R_{1}\right| \tag{2.16}
\end{equation*}
$$

Since $u \cdot v^{T}$ contains $\left|R_{1}\right| \cdot\left|C_{1}\right|$ 's, we can lower bound OPT $=\left\|A-u \cdot v^{T}\right\|_{0}$ by $\|A\|_{0}-\left|R_{1}\right| \cdot\left|C_{1}\right|$. Together with the assumption, we obtain that

$$
\mathrm{OPT} \geqslant\|A\|_{0}-\left|R_{1}\right| \cdot\left|C_{1}\right|>(1+\varepsilon) \mathrm{OPT}-\left|R_{1}\right| \cdot\left|C_{1}\right|,
$$

or equivalently, $\varepsilon \mathrm{OPT}<\left|R_{1}\right| \cdot\left|C_{1}\right|$. This implies OPT $/\left|R_{1}\right|<\left|C_{1}\right| / \varepsilon$ and thus by inequality (2.16) we have $\left|C_{0} \cap W\right| \leqslant 4\left|C_{1}\right| / \varepsilon$. Further, since $\left|C_{1} \cap W\right| \leqslant\left|C_{1}\right| \leqslant\left|C_{1}\right| / \varepsilon$ and $\left(C_{0} \cap W\right) \cup\left(C_{1} \cap W\right)=W$, we obtain the claimed upper bound of $|W| \leqslant 5\left|C_{1}\right| / \varepsilon$.

Claim 2.24. If $\mathrm{OPT}<\|A\|_{0} /(1+\varepsilon)$, then we have $\operatorname{Pr}[\mathcal{E}] \geqslant\left(\frac{\varepsilon}{5 t}\right)^{t}$.
Proof. If $\left|C_{1}\right| \geqslant t$ then event $\mathcal{E}$ asserts that $C_{1}^{\prime} \subseteq C_{1}$. Since $C_{1} \subseteq W$ (Claim 2.20) and $|W| \leqslant 5\left|C_{1}\right| / \varepsilon$ (Claim 2.23), a random element from $W$ hits $C_{1}$ with probability at least $\varepsilon / 5$. It follows that all $t$ samples forming $C_{1}^{\prime}$ hit $C_{1}$ with probability at least $(\varepsilon / 5)^{t}$.

If $\left|C_{1}\right|<t$ then event $\mathcal{E}$ asserts that $C_{1}^{\prime}=C_{1}$. Note that the $i$-th sample from $W$ is equal to the $i$-th element of $C_{1}$ with probability $1 /|W| \geqslant \varepsilon /\left(5\left|C_{1}\right|\right)>\varepsilon /(5 t)$. It follows that $C_{1}^{\prime}=C_{1}$ with probability at least $(\varepsilon /(5 t))^{\left|C_{1}\right|} \geqslant(\varepsilon /(5 t))^{t}$.

Lemma 2.25. Algorithm \& computes vectors $\widetilde{u} \in\{0,1\}^{m}$ and $\widetilde{v} \in\{0,1\}^{n}$ such that with probability at least $\left(\frac{\varepsilon}{5 t}\right)^{t+1}$ it holds that $\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0} \leqslant(1+\varepsilon) \mathrm{OPT}$.
Proof. When OPT $\geqslant\|A\|_{0} /(1+\varepsilon)$, for the basic solution $\left(0_{m}, 0_{n}\right)$ from Step 1 we have $\left\|A-0_{m} \cdot 0_{n}^{T}\right\|_{0}=$ $\|A\|_{0} \leqslant(1+\varepsilon) \mathrm{OPT}$. Otherwise, if OPT $<\|A\|_{0} /(1+\varepsilon)$, the statement follows by combining Claim 2.22 and Claim 2.24.

### 2.4.4 Proof of Theorem 1.3

By Lemma 2.25, Algorithm 4 computes vectors $\widetilde{u}, \widetilde{v}$ satisfying $\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0} \leqslant(1+\varepsilon)$ OPT with probability at least $\left(\frac{\varepsilon}{5 t}\right)^{t+1}$. By Lemma 2.19, the algorithm runs in time

$$
O\left(t^{2}\left(\|A\|_{0}+m+n\right) \log (m) \log (n)\right)
$$

Repeating the algorithm $O\left((5 t / \varepsilon)^{t+1} \log n\right)$ times and returning the best solution found, yields a success probability of $1-n^{-\Omega(1)}$. Since we set $t=2^{16} / \varepsilon^{2}$, the total running time is

$$
O\left((5 t / \varepsilon)^{t+3}\left(\|A\|_{0}+m+n\right) \log ^{3}(m n)\right)=(1 / \varepsilon)^{O\left(1 / \varepsilon^{2}\right)}\left(\|A\|_{0}+m+n\right) \log ^{3}(m n)
$$

## Chapter 3

## Algorithmic Results For Binary $\ell_{0}$-Rank-1 With Small Optimal Value

Given a binary matrix $A \in\{0,1\}^{m \times n}$ with $m \geqslant n$, our goal is to compute an approximate solution of the Binary $\ell_{0}$-Rank- 1 problem, and let us denote the optimal value by

$$
\begin{equation*}
\mathrm{OPT} \stackrel{\text { def }}{=} \min _{u \in\{0,1\}^{m}, v \in\{0,1\}^{n}}\left\|A-u \cdot v^{T}\right\|_{0} . \tag{3.1}
\end{equation*}
$$

In practice, approximating a matrix $A$ by a rank-1 matrix $u v^{T}$ makes most sense if $A$ is close to being rank-1. Hence, the above optimization problem is most relevant in the case when OPT $\ll\|A\|_{0}$. For this reason, we focus our studies on the setting when the ratio OPT/ $\|A\|_{0} \leqslant \phi$ for sufficiently small $\phi>0$.

Organization In Section 3.1, we give a simple $(1+O(\phi))$-approximation algorithm running in time $O\left(\min \left\{\|A\|_{0}+m+n, \phi^{-1}(m+n) \log (m n)\right\}\right)$. Then, in Section 3.2, we establish a sample complexity lower bound of $\Omega((m+n) / \phi)$ for any $(1+O(\phi))$-approximation algorithm, showing that our algorithm has an optimal runtime up to a poly $\log (m n)$ factor. In Section 3.3, we give an algorithm that runs in time $2^{O\left(\mathrm{OPT} / \sqrt{\|A\|_{0}}\right)} \cdot \operatorname{poly}(m n)$ and solves exactly the Binary $\ell_{0}$-Rank-1 problem.

### 3.1 A Simple Approximation Algorithm

We start by stating our core algorithmic result, which requires as an input a parameter $\phi \geqslant \mathrm{OPT} /\|A\|_{0}$. Theorem 1.4 (from page 7). Given $A \in\{0,1\}^{m \times n}$ with row and column sums, and given $\phi \in\left(0, \frac{1}{80}\right]$ with $\mathrm{OPT} /\|A\|_{0} \leqslant \phi$, we can compute in time $O\left(\min \left\{\|A\|_{0}+m+n, \phi^{-1}(m+n) \log (m n)\right\}\right)$ vectors $\widetilde{u} \in\{0,1\}^{m}$ and $\widetilde{v} \in\{0,1\}^{n}$ such that w.h.p. $\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0} \leqslant(1+5 \phi) \mathrm{OPT}+37 \phi^{2}\|A\|_{0}$.

In Chapter 4, see Section 4.2, we give a $(2+\varepsilon)$-approximation algorithm for the Reals $\ell_{0}$-Rank- 1 problem, which captures as a special case the Binary $\ell_{0}$-Rank-1 problem. In particular, we obtain a $(2+\varepsilon)$-approximation of OPT, and thus a $(2+\varepsilon)$-approximation of the ratio OPT/ $\|A\|_{0}$.
Theorem 1.5 (from page 7). Given $A \in\{0,1\}^{m \times n}$ with column adjacency arrays and OPT $\geqslant 1$, and given $\varepsilon \in(0,0.1]$, we can compute w.h.p. in time

$$
O\left(\left(\frac{n \log m}{\varepsilon^{2}}+\min \left\{\|A\|_{0}, n+\psi^{-1} \frac{\log n}{\varepsilon^{2}}\right\}\right) \frac{\log ^{2} n}{\varepsilon^{2}}\right)
$$

a column $A_{:, j}$ and a vector $z \in\{0,1\}^{n}$ such that w.h.p. $\left\|A-A_{:, j} \cdot z^{T}\right\|_{0} \leqslant(2+\varepsilon)$ OPT. Further, we can compute an estimate $Y$ such that w.h.p. $(1-\varepsilon) \mathrm{OPT} \leqslant Y \leqslant(2+2 \varepsilon) \mathrm{OPT}$.

Combining Theorem 1.4 and Theorem 1.5, yields an algorithm that does not need $\phi$ as an input and computes a $(1+500 \psi)$-approximate solution of the Binary $\ell_{0}$-Rank-1 problem.

Theorem 1.6 (from page 7). Given $A \in\{0,1\}^{m \times n}$ with column adjacency arrays and with row and column sums, for $\psi=\mathrm{OPT} /\|A\|_{0}$ we can compute w.h.p. in time $O\left(\min \left\{\|A\|_{0}+m+n, \psi^{-1}(m+n)\right\} \cdot \log ^{3}(m n)\right)$ vectors $\widetilde{u} \in\{0,1\}^{m}$ and $\widetilde{v} \in\{0,1\}^{n}$ such that w.h.p. $\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0} \leqslant(1+500 \psi)$ OPT.

Proof of Theorem 1.6. We compute a 3-approximation of OPT by applying Theorem 1.5 with $\varepsilon=0.1$. This yields a value $\phi$ satisfying $\psi \leqslant \phi \leqslant 3 \psi$. If $\phi>1 / 80$, then the 3 -approximation is already good enough, since $\psi>1 / 240$ and $1+500 \psi>3$. Otherwise, we run Theorem 1.4 with $\phi$. Further, using $\phi^{2}\|A\|_{0} \leqslant 9 \psi^{2}\|A\|_{0}=9 \psi \mathrm{OPT}$, the total error is at most

$$
(1+5 \phi) \mathrm{OPT}+37 \phi^{2}\|A\|_{0} \leqslant(1+15 \psi) \mathrm{OPT}+37 \cdot 9 \psi \mathrm{OPT} \leqslant(1+500 \psi) \mathrm{OPT} .
$$

A rough upper bound on the running time is $O\left(\min \left\{\|A\|_{0}+m+n, \psi^{-1}(m+n)\right\} \cdot \log ^{3}(m n)\right)$.

The remainder of this section is devoted to proving Theorem 1.4.

### 3.1.1 Preparations

Given a matrix $A \in\{0,1\}^{m \times n}$, let $u \in\{0,1\}^{m}$ and $v \in\{0,1\}^{n}$ be an optimal solution to the Binary $\ell_{0}$-Rank-1 problem, realizing OPT $=\left\|A-u \cdot v^{T}\right\|_{0}$. Moreover, set $\alpha \stackrel{\text { def }}{=}\|u\|_{0}$ and $\beta \stackrel{\text { def }}{=}\|v\|_{0}$. We start with the following technical preparations.

Lemma 3.1. For any row $i \in\{1, \ldots, m\}$ let $x_{i}$ be the number of 0 's in columns selected by $v$, i.e., $x_{i} \stackrel{\text { def }}{=}\left\{j \in[n]: A_{i, j}=0, v_{j}=1\right\}$, and let $y_{i}$ be the number of 1's in columns not selected by $v$, i.e., $y_{i} \stackrel{\text { def }}{=}\left\{j \in[n]: A_{i, j}=1, v_{j}=0\right\}$. Let $R=\left\{i \in[m]: u_{i}=1\right\}$ be the rows selected by $u$, and let $\bar{R} \stackrel{\text { def }}{=}[m] \backslash R$. Symmetrically, let $C$ be the columns selected by $v$. Then we have

1. $\left\|A_{i,}:\right\|_{0}=\beta-x_{i}+y_{i}$, for any $i \in[m]$;
2. $\mathrm{OPT}=\sum_{i \in R}\left(x_{i}+y_{i}\right)+\sum_{i \in \bar{R}}\left(\beta-x_{i}+y_{i}\right)$;
3. $\mathrm{OPT} \geqslant \sum_{i \in R}\left|x_{i}-y_{i}\right|$;
4. $x_{i} \leqslant \beta / 2$ for any $i \in R$, and $x_{i} \geqslant \beta / 2$, for any $i \in \bar{R}$;
5. OPT $\geqslant \sum_{i=1}^{m} \min \left\{\left\|A_{i,}:\right\|_{0},\left|\left\|A_{i,:}\right\|_{0}-\beta\right|\right\}$;
6. $\left|\left|\left|A \|_{0}-\alpha \beta\right| \leqslant \mathrm{OPT}\right.\right.$;
7. If $\mathrm{OPT} \leqslant \phi\|A\|_{0}$, then $(1-\phi)\|A\|_{0} \leqslant \alpha \beta \leqslant(1+\phi)\|A\|_{0}$.

Proof. For (1), note that in the $\beta$ columns $C$ selected by $v$, row $i$ has $\beta-x_{i} 1$ 's, and in the remaining $n-|C|$ columns row $i$ has $y_{i} 1$ 's. Hence, the total number of 1 's in row $i$ is $\left\|A_{i,:}\right\|_{0}=\beta-x_{i}+y_{i}$.
(2) We split $\mathrm{OPT}=\left\|A-u v^{T}\right\|_{0}$ into a sum over all rows, so that $\mathrm{OPT}=\sum_{i=1}^{m}\left\|A_{i,:}-u_{i} v^{T}\right\|_{0}$. For $i \in \bar{R}$, the $i$-th term of this sum is simply $\left\|A_{i,}:\right\|_{0}=\beta-x_{i}+y_{i}$. For $i \in R$, the $i$-th term is $\left\|A_{i,:}-v^{T}\right\|_{0}=x_{i}+y_{i}$.
(3) follows immediately from (2).
(4) follows from (2), since for $x_{i}>\beta / 2$ and $i \in R$ we can change $u_{i}$ from 1 to 0 , reducing the contribution of row $i$ from $x_{i}+y_{i}$ to $\beta-x_{i}+y_{i}$, which contradicts optimality of OPT.

For (5), we use that $x_{i}+y_{i} \geqslant\left|x_{i}-y_{i}\right|=\left|\left|\left|A_{i,}: \|_{0}-\beta\right|\right.\right.$ by (1), and

$$
\mathrm{OPT}=\sum_{i \in R}\left(x_{i}+y_{i}\right)+\sum_{i \in \bar{R}}\left(\beta-x_{i}+y_{i}\right)=\sum_{i \in R}\left(x_{i}+y_{i}\right)+\sum_{i \in \bar{R}}\left\|A_{i,:}\right\|_{0} .
$$

(6) is shown similarly to (5) by noting that

$$
\begin{aligned}
\mathrm{OPT} & =\sum_{i \in R}\left(x_{i}+y_{i}\right)+\sum_{i \in \bar{R}}\left(\beta-x_{i}+y_{i}\right) \geqslant \sum_{i \in R}\left|\left\|A_{i,:}\right\|_{0}-\beta\right|+\sum_{i \in \bar{R}}\left\|A_{i,:}\right\|_{0} \\
& \geqslant \sum_{i \in R}\left(\left\|A_{i,:}\right\|_{0}-\beta\right)+\sum_{i \in \bar{R}}\left\|A_{i,:}\right\|_{0}=\|A\|_{0}-\alpha \beta
\end{aligned}
$$

and similarly

$$
\mathrm{OPT} \geqslant \sum_{i \in R}\left|\left\|A_{i,:}\right\|_{0}-\beta\right|+\sum_{i \in \bar{R}}\left\|A_{i,:}:\right\|_{0} \geqslant \sum_{i \in R}\left(\beta-\left\|A_{i,:}:\right\|_{0}\right)-\sum_{i \in \bar{R}}\left\|A_{i,:}\right\|_{0}=\alpha \beta-\|A\|_{0}
$$

Finally, (7) follows immediately from (6) by plugging in the upper bound OPT $\leqslant \phi\|A\|_{0}$.

### 3.1.2 Approximating $\alpha$ and $\boldsymbol{\beta}$

We now show how to approximate $\alpha=\|u\|_{0}$ and $\beta=\|v\|_{0}$, where $(u, v)$ is an optimal solution.
Lemma 3.2. Given $A \in\{0,1\}^{m \times n}$ and $\phi \in(0,1 / 30]$ with $\mathrm{OPT} /\|A\|_{0} \leqslant \phi$, we can compute in time $O\left(\|A\|_{0}+m+n\right)$ an integer $\widetilde{\beta} \in[m]$ with

$$
\frac{1-3 \phi}{1-\phi} \beta \leqslant \widetilde{\beta} \leqslant \frac{1+\phi}{1-\phi} \beta .
$$

Symmetrically, we can approximate $\alpha$ by $\widetilde{\alpha}$. If we are additionally given the number of 1's in each row and column, then the running time becomes $O(m+n)$.

Proof. Let

$$
\Lambda \stackrel{\text { def }}{=} \min _{\beta^{\prime} \in[n]} \sum_{i=1}^{m} \min \left\{\left\|A_{i,:}\right\|_{0},\left|\left\|A_{i,:}\right\|_{0}-\beta^{\prime}\right|\right\}
$$

and let $\widetilde{\beta}$ be the value of $\beta^{\prime}$ realizing $\Lambda$.
We first verify the approximation guarantee. Consider the set of rows $R$ selected by $u$. Let $x_{i}, y_{i}$ for $i \in R$ be as in Lemma 3.1. Then we have

$$
\Lambda \geqslant \sum_{i \in R} \min \left\{\left\|A_{i,:}\right\|_{0},\left|\left\|A_{i,:}\right\|_{0}-\widetilde{\beta}\right|\right\}=\sum_{i \in R} \min \left\{\beta+y_{i}-x_{i},\left|\beta-\widetilde{\beta}+y_{i}-x_{i}\right|\right\},
$$

where we used Lemma 3.1.1. Assume for the sake of contradiction that $|\beta-\widetilde{\beta}|>\frac{2 \phi}{1-\phi} \beta$. Since $|x-y| \geqslant$ $|x|-|y|$ for any numbers $x, y$, we obtain

$$
\left|\beta-\widetilde{\beta}+y_{i}-x_{i}\right| \geqslant|\beta-\widetilde{\beta}|-\left|x_{i}-y_{i}\right|>\frac{2 \phi}{1-\phi} \beta-\left|x_{i}-y_{i}\right| .
$$

Similarly, we have $\beta+y_{i}-x_{i}>\frac{2 \phi}{1-\phi} \beta-\left|x_{i}-y_{i}\right|$. Hence,

$$
\Lambda>\sum_{i \in R}\left(\frac{2 \phi \beta}{1-\phi}-\left|x_{i}-y_{i}\right|\right) \geqslant \frac{2 \phi \beta}{1-\phi}|R|-\mathrm{OPT}
$$

where we used Lemma 3.1.3. Since $R$ is the set of rows selected by $u$, we have $|R|=\alpha$. By Lemma 3.1.7, we have OPT $\leqslant \phi\|A\|_{0} \leqslant \frac{\phi}{1-\phi} \alpha \beta$. Together, this yields $\Lambda>$ OPT, contradicting

$$
\Lambda \leqslant \sum_{i=1}^{m} \min \left\{\left\|A_{i,:}\right\|_{0},\left|\left|A_{i,:} \|_{0}-\beta\right|\right\} \leqslant \mathrm{OPT}\right.
$$

by Lemma 3.1.5. Hence, $|\beta-\widetilde{\beta}| \leqslant \frac{2 \phi}{1-\phi} \beta$.
It remains to design a fast algorithm. We first compute all numbers $\left\|A_{i,}:\right\|_{0}$ in time $O\left(\|A\|_{0}\right)$ (this step can be skipped if we are given these numbers as input). We sort these numbers, obtaining a sorted order $\left\|A_{\pi(1),:}\right\|_{0} \leqslant \ldots \leqslant\left\|A_{\pi(m),:}\right\|_{0}$. Using counting sort, this takes time $O(m+n)$. We precompute prefix sums $P(k) \stackrel{\text { def }}{=} \sum_{\ell=1}^{k}\left\|A_{\pi(\ell) \text { : }}\right\|_{0}$, which allows us to evaluate in constant time any sum

$$
\sum_{\ell=x}^{y}\left\|A_{\pi(\ell),:}\right\|_{0}=P(y)-P(x-1) .
$$

Finally, we precompute the inverse

$$
\ell\left(\beta^{\prime}\right) \stackrel{\text { def }}{=} \max \left\{\ell:\left\|A_{\pi(\ell),:}\right\|_{0} \leqslant \beta^{\prime}\right\}
$$

or $\ell\left(\beta^{\prime}\right)=0$ if there is no $\ell$ with $\left\|A_{\pi(\ell) \text {, }}\right\|_{0} \leqslant \beta^{\prime}$. By a simple sweep, all values $\ell\left(\beta^{\prime}\right)$ can be computed in total time $O(m+n)$.

Note that for any fixed $\beta^{\prime}$ and row $i$, the term realizing $\min \left\{\left\|A_{i,:}\right\|\left\|_{0},\left|\left\|A_{i,:}\right\|_{0}-\beta^{\prime}\right|\right\}\right.$ is equal to:
(a) $\left\|A_{i,:}\right\|_{0}$ if $\left\|A_{i,:}\right\|_{0} \leqslant \beta^{\prime} / 2$;
(b) $\beta^{\prime}-\left\|A_{i,:}\right\|_{0}$, if $\beta^{\prime} / 2<\left\|A_{i,:}\right\|_{0} \leqslant \beta^{\prime} ; \quad$ and
(c) $\left\|A_{i,:}\right\|_{0}-\beta^{\prime}$, if $\left\|A_{i,:}\right\|_{0}>\beta^{\prime}$. Hence, we obtain

$$
\begin{aligned}
& \sum_{i=1}^{n} \min \left\{\left\|A_{i,:}\right\|_{0},\left|\left\|A_{i,:}:\right\|_{0}-\beta^{\prime}\right|\right\} \\
& =\left(\sum_{i=1}^{\ell\left(\beta^{\prime} / 2\right)}\left\|A_{\pi(i),:}\right\|_{0}\right)+\left(\sum_{i=\ell\left(\beta^{\prime} / 2\right)+1}^{\ell\left(\beta^{\prime}\right)} \beta^{\prime}-\left\|A_{\pi(i),:}\right\|_{0}\right)+\left(\sum_{i=\ell\left(\beta^{\prime}\right)+1}^{n}\left\|A_{\pi(i),:}\right\|_{0}-\beta^{\prime}\right) \\
& =P(n)-2\left[P\left(\ell\left(\beta^{\prime}\right)\right)-P\left(\ell\left(\beta^{\prime} / 2\right)\right)\right]-\left[n+\ell\left(\beta^{\prime} / 2\right)-2 \cdot \ell\left(\beta^{\prime}\right)\right] \beta^{\prime} .
\end{aligned}
$$

This shows that after the above precomputation the $\operatorname{sum} \sum_{j=1}^{n} \min \left\{\left\|A_{i,:}:\right\|,\left|\left\|A_{: j}\right\|_{0}-\beta^{\prime}\right|\right\}$ can be evaluated in time $O(1)$ for any $\beta^{\prime}$. Minimizing over all $\beta^{\prime} \in[m]$ yields $\widetilde{\beta}$. This finishes our algorithm, which runs in total time $O\left(\|A\|_{0}+m+n\right)$, or $O(m+n)$ if we are given the number of 1 's in each row and column.

### 3.1.3 The Algorithm

We now design an approximation algorithm for the Binary $\ell_{0}$-Rank- 1 problem that will yield Theorem 1.4. We present the pseudocode of this Algorithm 5 below.

```
Algorithm 5 Binary \(\ell_{0}\)-Rank-1 With Small Optimal Value
Input: \(A \in\{0,1\}^{m \times n}\) and \(\phi \in(0,1 / 80]\) such that OPT/ \(\|A\|_{0} \leqslant \phi\).
Output: Vectors \(\widetilde{u} \in\{0,1\}^{m}, \widetilde{v} \in\{0,1\}^{n}\) such that \(\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0} \leqslant(1+\varepsilon)\) OPT.
```

1. Compute approximations $\frac{1-3 \phi}{1-\phi} \alpha \leqslant \widetilde{\alpha} \leqslant \frac{1+\phi}{1-\phi} \alpha$ and $\frac{1-3 \phi}{1-\phi} \beta \leqslant \widetilde{\beta} \leqslant \frac{1+\phi}{1-\phi} \beta$ using Lemma 3.2.

Initialize $R^{R} \stackrel{\text { def }}{=}[m], C^{R} \stackrel{\text { def }}{=}[n], R^{S} \stackrel{\text { def }}{=} \emptyset, C^{S} \stackrel{\text { def }}{=} \emptyset$.
2. For any row $i$, if $\left\|A_{i,}:\right\|_{0}<\frac{1-\phi}{1+\phi} \cdot \frac{\widetilde{\beta}}{2}$ then set $\widetilde{u}_{i}=0$ and remove $i$ from $R^{R}$.

For any column $j$, if $\left\|A_{:, j}\right\|_{0}<\frac{1-\phi}{1+\phi} \cdot \frac{\widetilde{\alpha}}{2}$ then set $\widetilde{v}_{j}=0$ and remove $j$ from $C^{R}$.
3. For any $i \in R^{R}$ compute an estimate $X_{i}$ with $\left|X_{i}-\left\|A_{i, C^{R}}\right\|_{0}\right| \leqslant \frac{1}{9}\left|C^{R}\right|$.

For any $j \in C^{R}$ compute an estimate $Y_{j}$ with $\left|Y_{j}-\left\|A_{R^{R}, j}\right\|_{0}\right| \leqslant \frac{1}{9}\left|R^{R}\right|$.
4. For any $i \in R^{R}$, if $X_{i}>\frac{2}{3} \widetilde{\beta}$ then set $\widetilde{u}_{i}=1$ and add $i$ to $R^{S}$.

For any $j \in C^{R}$, if $Y_{j}>\frac{2}{3} \widetilde{\alpha}$ then set $\widetilde{v}_{j}=1$ and add $j$ to $C^{S}$.
5. For any $i \in R^{R} \backslash R^{S}$, compute an estimate $X_{i}^{\prime}$ with $\left|X_{i}^{\prime}-\left\|A_{i, C^{S}}\right\|_{0}\right| \leqslant \phi\left|C^{S}\right|$, For any $j \in C^{R} \backslash C^{S}$, compute an estimate $Y_{j}^{\prime}$ with $\left|Y_{j}^{\prime}-\left\|A_{R^{S}, j}\right\|_{0}\right| \leqslant \phi\left|R^{S}\right|$.
6. For any $i \in R^{R} \backslash R^{S}$, set $\widetilde{u}_{i}=1$ if $X_{i}^{\prime} \geqslant\left|C^{S}\right| / 2$ and 0 otherwise,

For any $j \in C^{R} \backslash C^{S}$, set $\widetilde{v}_{j}=1$ if $Y_{j}^{\prime} \geqslant\left|R^{S}\right| / 2$ and 0 otherwise.
7. Return $(\widetilde{u}, \widetilde{v})$.

Running Time By Lemma 3.2, Step 1 runs in time $O\left(\|A\|_{0}+m+n\right)$, or in time $O(m+n)$ if we are given the number of 1's in each row and column. Steps 2, 4, and 6 clearly run in time $O(m+n)$. For steps 3 and 5 , there are two ways to implement them.

Variant (1) is an exact algorithm. We enumerate all nonzero entries of $A$ and count how many contribute to the required numbers $\left\|A_{i, C^{R}}\right\|_{0},\left\|A_{R^{R}, j}\right\|_{0}$ etc. This takes total time $O\left(\|A\|_{0}\right)$, and hence the total running time of the algorithm is $O\left(\|A\|_{0}+m+n\right)$.

Variant (2) uses random sampling. In order to estimate $\left\|A_{i, C^{R}}\right\|_{0}$, consider a random variable $Z$ that draws a uniformly random column $j \in C^{R}$ and returns 1 if $A_{i, j} \neq 0$ and 0 otherwise. Then $\mathbb{E}[Z]=\left\|A_{i, C^{R}}\right\|_{0} /\left|C^{R}\right|$. Taking independent copies $Z_{1}, \ldots, Z_{\ell}$ of $Z$, where $\ell=\Theta\left(\log (m n) / \delta^{2}\right)$ with sufficiently large hidden constant, a standard Chernoff bound argument shows that w.h.p.

$$
\left|\left(Z_{1}+\ldots+Z_{\ell}\right) \cdot \frac{\left|C^{R}\right|}{\ell}-\left\|A_{i, C^{R}}\right\|_{0}\right| \leqslant \delta \cdot\left|C^{R}\right|
$$

which yields the required approximation. For Step 3 we use this procedure with $\delta=\frac{1}{9}$ and obtain running time $O(\log (m n))$ per row and column, or $O((m+n) \log (m n))$ in total. For Step 5 we use $\delta=\phi$, resulting in time $O\left(\phi^{-2} \log (m n)\right)$ for computing one estimate $X_{i}^{\prime}$ or $Y_{j}^{\prime}$. By Claim 3.6 below there are only $O(\phi(m+n))$ rows and columns in $R^{R} \backslash R^{S}$ and $C^{R} \backslash C^{S}$, and hence the total running time for Step 5 is $O\left(\phi^{-1}(m+n) \log (m n)\right)$. This dominates the total running time.

Combining both variants, we obtain the claimed running time of

$$
O\left(\min \left\{\|A\|_{0}+m+n, \phi^{-1}(m+n) \log (m n)\right\}\right)
$$

Correctness In the following we analyze the correctness of the above algorithm.
Claim 3.3. For any row $i$ deleted in Step 2 we have $\widetilde{u}_{i}=u_{i}$. Symmetrically, for any column $j$ deleted in Step 2 we have $\widetilde{v}_{j}=v_{j}$.
Proof. If row $i$ is deleted, then by the approximation guarantee of $\widetilde{\beta}$ we have

$$
\left\|A_{i,:}\right\|_{0}<\frac{1-\phi}{1+\phi} \cdot \frac{\widetilde{\beta}}{2} \leqslant \frac{\beta}{2}
$$

Note that for $x_{i}$ (the number of 0's in row $i$ in columns selected by $v$ ) we have $x_{i} \geqslant \beta-\left\|A_{i,:}\right\|_{0}$. Together, we obtain $x_{i}>\beta / 2$, and thus row $i$ cannot be selected by $u$, by Lemma 3.1.4. Hence, we have $u_{i}=0=\widetilde{u}_{i}$. The statement for the columns is symmetric.

Claim 3.4. After Step 2, it holds for the remaining rows $R^{R}$ and columns $C^{R}$ that

$$
\left|R^{R}\right| \leqslant\left(1+\frac{1+\phi}{1-3 \phi} \cdot \frac{2 \phi}{1-\phi}\right) \alpha \quad \text { and } \quad\left|C^{R}\right| \leqslant\left(1+\frac{1+\phi}{1-3 \phi} \cdot \frac{2 \phi}{1-\phi}\right) \beta
$$

Proof. By Claim 3.3 the $\alpha$ rows $R$ selected by $u$ remain. We split the rows $R^{R}$ remaining after Step 2 into $R \cup R^{\prime}$, and bound $\left|R^{\prime}\right|$ from above. Since any $i \in R^{\prime}$ is not selected by $u$, it contributes $\left\|A_{i,:}\right\|_{0}$ to OPT. Note that

$$
\left\|A_{i,:}\right\|_{0} \geqslant \frac{1-\phi}{1+\phi} \cdot \frac{\widetilde{\beta}}{2} \geqslant \frac{1-\phi}{1+\phi} \cdot \frac{1-3 \phi}{1-\phi} \cdot \frac{\beta}{2}=\frac{1-3 \phi}{1+\phi} \cdot \frac{\beta}{2}
$$

and thus $\left|R^{\prime}\right| \leqslant \mathrm{OPT} \cdot \frac{1+\phi}{1-3 \phi} \cdot \frac{2}{\beta}$. Since

$$
\mathrm{OPT} \leqslant \phi\|A\|_{0} \leqslant \frac{\phi}{1-\phi} \cdot \alpha \beta
$$

by Lemma 3.1.7, we obtain $\left|R^{\prime}\right| \leqslant \frac{1+\phi}{1-3 \phi} \cdot \frac{2 \phi}{1-\phi} \cdot \alpha$. Thus, we have in total

$$
\left|R^{R}\right|=|R|+\left|R^{\prime}\right| \leqslant\left(1+\frac{1+\phi}{1-3 \phi} \cdot \frac{2 \phi}{1-\phi}\right) \alpha
$$

The statement for the columns is symmetric.
Claim 3.5. The rows and columns selected in Step 4 are also selected by the optimal solution $u, v$, i.e., for any $i \in R^{S}$ we have $u_{i}=1$ and for any $j \in C^{S}$ we have $v_{j}=1$.

Proof. If row $i$ is selected in Step 4, then we have by the approximation guarantee of $X_{i}$, definition of Step 4, Claim 3.4, and Lemma 3.2

$$
\begin{aligned}
\left\|A_{i, C^{R}}\right\|_{0} & \geqslant X_{i}-\frac{1}{9}\left|C^{R}\right|>\frac{2}{3} \widetilde{\beta}-\frac{1}{9}\left(1+\frac{1+\phi}{1-3 \phi} \cdot \frac{2 \phi}{1-\phi}\right) \beta \\
& \geqslant \frac{2}{3} \cdot \frac{1-3 \phi}{1-\phi} \beta-\frac{1}{9}\left(1+\frac{1+\phi}{1-3 \phi} \cdot \frac{2 \phi}{1-\phi}\right) \beta
\end{aligned}
$$

It is easy to see that for sufficiently small $\phi \geqslant 0$ this yields

$$
\left\|A_{i, C^{R}}\right\|_{0}>\frac{\beta}{2}+\frac{1+\phi}{1-3 \phi} \cdot \frac{2 \phi}{1-\phi} \beta
$$

One can check that $0 \leqslant \phi \leqslant 1 / 80$ is sufficient. Since there are $\left|C^{R}\right| \leqslant\left(1+\frac{1+\phi}{1-3 \phi} \cdot \frac{2 \phi}{1-\phi}\right) \beta$ columns remaining, in particular the $\beta$ columns $C \subseteq C^{R}$ which are selected by $v$, we obtain

$$
\left\|A_{i, C}\right\|_{0} \geqslant\left\|A_{i, C^{R}}\right\|_{0}-\left(\left|C^{R}\right|-\beta\right)>\beta / 2
$$

By Lemma 3.1.4, we thus obtain that row $i$ is selected by the optimal $u$. The statement for the columns is symmetric.

Claim 3.6. After Step 4 there are $\left|R^{R} \backslash R^{S}\right| \leqslant 6 \phi \alpha$ remaining rows and $\left|C^{R} \backslash C^{S}\right| \leqslant 6 \phi \beta$ remaining columns.

Proof. After Step 4, every remaining row $i$, for any $0 \leqslant \phi \leqslant 1 / 80$, satisfies

$$
\left\|A_{i,:}\right\|_{0} \geqslant \frac{1-\phi}{1+\phi} \cdot \frac{\widetilde{\beta}}{2} \geqslant \frac{1-\phi}{1+\phi} \cdot \frac{1-3 \phi}{1-\phi} \cdot \frac{\beta}{2} \geqslant \frac{2}{5} \beta
$$

Moreover, each such row satisfies

$$
\left\|A_{i, C^{R}}\right\|_{0} \leqslant X_{i}+\frac{1}{9}\left|C^{R}\right| \leqslant \frac{2}{3} \widetilde{\beta}+\frac{1}{9}\left|C^{R}\right|
$$

which together with $\widetilde{\beta} \leqslant \frac{1+\phi}{1-\phi} \beta$ (Lemma 3.2) and $\left|C^{R}\right| \leqslant\left(1+\frac{1+\phi}{1-3 \phi} \cdot \frac{2 \phi}{1-\phi}\right) \beta$ (Claim 3.4) yields

$$
\left\|A_{i, C^{R}}\right\|_{0} \leqslant\left(\frac{2}{3} \cdot \frac{1+\phi}{1-\phi}+\frac{1}{9}+\frac{1}{9} \cdot \frac{1+\phi}{1-3 \phi} \cdot \frac{2 \phi}{1-\phi}\right) \beta .
$$

It is easy to see that for sufficiently small $\phi \geqslant 0$ we have $\left\|A_{i, C^{R}}\right\|_{0} \leqslant \frac{4}{5} \beta$, and it can be checked that $0 \leqslant \phi \leqslant 1 / 80$ is sufficient.

If $i$ is not selected by $u$, then its contribution to OPT is $\left\|A_{i,:}\right\|_{0} \geqslant \frac{2}{5} \beta$. If $i$ is selected by $u$, then since $C \subseteq C^{R}$ its contribution to OPT is at least

$$
\beta-\left\|A_{i, C}\right\|_{0} \geqslant \beta-\left\|A_{i, C^{R}}\right\|_{0} \geqslant \beta-\frac{4}{5} \beta=\frac{1}{5} \beta .
$$

Thus, the number of remaining rows is at most

$$
\frac{\mathrm{OPT}}{\beta / 5} \leqslant \frac{5 \phi \alpha \beta}{(1-\phi) \beta} \leqslant 6 \phi \alpha,
$$

where we used Lemma 3.1.7. The statement for the columns is symmetric.
We are now ready to prove correctness of Algorithm 5.
Proof of Theorem 1.4. The rows and columns removed in Step 2 are also not picked by the optimal solution, by Claim 3.3. Hence, in the region $\left([m] \backslash R^{R}\right) \times[n]$ and $[m] \times\left([n] \backslash C^{R}\right)$ we incur the same error as the optimal solution. The rows and columns chosen in Step 4 are also picked by the optimal solution, by Claim 3.5. Hence, in the region $R^{S} \times C^{S}$ we incur the same error as the optimal solution. We split the remaining matrix into three regions: $\left(R^{R} \backslash R^{S}\right) \times C^{S}, R^{S} \times\left(C^{R} \backslash C^{S}\right)$, and $\left(R^{R} \backslash R^{S}\right) \times\left(C^{R} \backslash C^{S}\right)$.

In the region $\left(R^{R} \backslash R^{S}\right) \times C^{S}$ we compute for any row $i \in R^{R} \backslash R^{S}$ an additive $\phi|C|$-approximation $X_{i}^{\prime}$ of $\left\|A_{i, C}\right\|_{0}$, and we pick row $i$ iff $X_{i}^{\prime} \geqslant|C| / 2$. In case $\left|\left\|A_{i, C}\right\|_{0}-|C| / 2\right|>\phi|C|$, we have $X_{i}^{\prime} \geqslant|C| / 2$ if and only if $\left\|A_{i, C}\right\|_{0} \geqslant|C| / 2$, and thus our choice for row $i$ is optimal, restricted to region $\left(R^{R} \backslash R^{S}\right) \times C^{S}$. Otherwise, if $\left|\left\|A_{i, C}\right\|_{0}-|C| / 2\right| \leqslant \phi|C|$, then no matter whether we choose row $i$ or not, we obtain approximation ratio

$$
\frac{|C| / 2+\phi|C|}{|C| / 2-\phi|C|}=\frac{1+2 \phi}{1-2 \phi} \leqslant 1+5 \phi,
$$

restricted to region $\left(R^{R} \backslash R^{S}\right) \times C^{S}$. The region $R^{S} \times\left(C^{R} \backslash C^{S}\right)$ is symmetric.
Finally, in region $\left(R^{R} \backslash R^{S}\right) \times\left(C^{R} \backslash C^{S}\right)$ we pessimistically assume that every entry is an error. By Claim 3.6 and Lemma 3.1.7, this submatrix has size at most

$$
6 \phi \alpha \cdot 6 \phi \beta \leqslant 36 \phi^{2}(1+\phi)\|A\|_{0} \leqslant 37 \phi^{2}\|A\|_{0} .
$$

In total, over all regions, we computed vectors $\widetilde{u}, \widetilde{v}$ such that

$$
\left\|A-\widetilde{u} \widetilde{v}^{T}\right\|_{0} \leqslant(1+5 \phi) \mathrm{OPT}+37 \phi^{2}\|A\|_{0}
$$

This completes the correctness prove of Algorithm 5.

### 3.2 Sample Complexity Lower Bound

We give now a lower bound of $\Omega(n / \phi)$ on the number of samples of any $1+O(\phi)$-approximation algorithm for the Binary $\ell_{0}$-Rank- 1 problem, where $\phi \geqslant \mathrm{OPT} /\|A\|_{0}$ as before.

Theorem 1.7 (from page 7). Let $C \geqslant 1$. Given an $n \times n$ binary matrix $A$ with column adjacency arrays and with row and column sums, and given $\sqrt{\log (n) / n} \ll \phi \leqslant 1 / 100 C$ such that OPT $\|A\|_{0} \leqslant \phi$, computing a (1+C $)$-approximation of OPT requires to read $\Omega(n / \phi)$ entries of $A$ (in the worst case over A).

The technical core of our argument is the following lemma.
Lemma 3.7. Let $\phi \in(0,1 / 2)$. Let $X_{1}, \ldots, X_{k}$ be binary random variables with expectations $p_{1}, \ldots, p_{k}$, where $p_{i} \in\{1 / 2-\phi, 1 / 2+\phi\}$ for each $i$. Let $\mathcal{A}$ be an algorithm which can adaptively obtain any number of samples of each random variable, and which outputs bits $b_{i}$ for every $i \in[1: k]$. Suppose that with probability at least 0.95 over the joint probability space of $\mathcal{A}$ and the random samples, $\mathcal{A}$ outputs for at least a 0.95 fraction of all $i$ that $b_{i}=1$ if $p_{i}=1 / 2+\phi$ and $b_{i}=0$ otherwise. Then, with probability at least $0.05, \mathcal{A}$ makes $\Omega\left(k / \phi^{2}\right)$ samples in total, asymptotically in $k$.

Proof. Consider the following problem $P$ : let $X$ be a binary random variable with expectation $p$ drawn uniformly in $\{1 / 2-\phi, 1 / 2+\phi\}$. It is well-known that any algorithm which, with probability at least 0.6 , obtains samples from $X$ and outputs 0 if $p=1 / 2-\phi$ and outputs 1 if $p=1 / 2+\phi$, requires $\Omega\left(1 / \phi^{2}\right)$ samples; see, e.g., Theorem 4.32 of [BY02]. Let $c>0$ be such that $c / \phi^{2}$ is a lower bound on the number of samples for this problem $P$.

Let $\mathcal{A}$ be an algorithm solving the problem in the lemma statement. Since $\mathcal{A}$ succeeds with probability at least 0.95 in obtaining the guarantees of the lemma for given sequence $p_{1}, \ldots, p_{k}$, it also succeeds with this probability when $\left(p_{1}, \ldots, p_{k}\right)$ is drawn from the uniform distribution on $\{1 / 2-\phi, 1 / 2+\phi\}^{k}$.

Suppose, towards a contradiction, that $\mathcal{A}$ takes less than $0.05 \cdot c k / \phi^{2}$ samples with probability at least 0.95 . By stopping $\mathcal{A}$ before taking $0.05 \cdot c k / \phi^{2}$ samples, we obtain an algorithm $A^{\prime}$ that always takes less than $0.05 \cdot c k / \phi^{2}$ samples. By the union bound, $A^{\prime}$ obtains the guarantees of the lemma for the output bits $b_{i}$ with probability at least 0.9 , over the joint probability space of $A^{\prime}$ and the random samples.

Note that the expected number of samples $A^{\prime}$ takes from a given $X_{i}$ is less than $0.05 \cdot c / \phi^{2}$. By Markov's inequality, for a 0.95 fraction of indices $i, A^{\prime}$ takes less than $c / \phi^{2}$ samples from $X_{i}$. We say that $i$ is good if $A^{\prime}$ takes less than $c / \phi^{2}$ samples from $X_{i}$ and the output bit $b_{i}$ is correct. By union bound, at least a $1-(1-0.9)-(1-0.95)=0.85$ fraction of indices $i$ is good.

Since $\left(p_{1}, \ldots, p_{k}\right)$ is drawn from the uniform distribution on $\{1 / 2-\phi, 1 / 2+\phi\}^{k}$, with probability at least 0.95 the number $k_{+}=\left|\left\{i: p_{i}=1 / 2+\phi\right\}\right|$ satisfies $0.45 k \leqslant k_{+} \leqslant 0.55 k$ (for sufficiently large $k$ ). This implies that a 0.65 fraction of indices $\left\{i: p_{i}=1 / 2+\phi\right\}$ is good, as otherwise the number of bad $i$ 's is at least $(1-0.65) \cdot 0.45 k>0.15 k$. Similarly, a 0.65 fraction of indices $\left\{i: p_{i}=1 / 2-\phi\right\}$ is good.

Given an instance of problem $P$ with random variable $X$ and expectation $p$, we choose a uniformly random $i \in[k]$, and set $X_{i}=X$. For $j \neq i$, we independently and uniformly at random choose $p_{j} \in$ $\{1 / 2-\phi, 1 / 2+\phi\}$. We then run algorithm $A^{\prime}$. Whenever $A^{\prime}$ samples from $X_{i}$, we sample a new value of $X$ as in problem $P$. Whenever $A^{\prime}$ samples from $X_{j}$ for $j \neq i$, we flip a coin with probability $p_{j}$ and report the output to $A^{\prime}$. If $A^{\prime}$ takes $c / \phi^{2}$ samples from $X_{i}$, then we abort, thus ensuring that $A^{\prime}$ always takes less than $c / \phi^{2}$ samples from $X_{i}=X$. Observe that the input to $A^{\prime}$ is a sequence of random variables $X_{1}, \ldots, X_{k}$ with expectations $p_{1}, \ldots, p_{k}$ which are independent and uniformly distributed in $\{1 / 2-\phi, 1 / 2+\phi\}$. In particular, except for their expectation these random variables are indistinguishable.

We now condition on $0.45 k \leqslant k_{+} \leqslant 0.55 k$, which has success probability at least 0.95 for sufficiently large $k$. Then no matter whether $p_{i}=1 / 2+\phi$ or $p_{i}=1 / 2-\phi$, at least a 0.65 fraction of indices $j$ with $p_{j}=p_{i}$ is good. Since $i$ was chosen to be a uniformly random position independently of the randomness of the sampling and the algorithm $A^{\prime}$, and the $X_{j}$ with $p_{j}=p_{i}$ are indistinguishable, with probability at least 0.65 index $i$ is good. In this case, $A^{\prime}$ takes less than $c / \phi^{2}$ samples from $X_{i}=X$ and correctly determines the output bit $b_{i}$, i.e., whether $p_{i}=1 / 2+\phi$. As by union bound the total success probability is $1-(1-0.65)-(1-0.95)=0.6$, this contradicts the requirement of $c / \phi^{2}$ samples mentioned above for solving $P$. Hence, the assumption was wrong, and $\mathcal{A}$ takes $\Omega\left(k / \phi^{2}\right)$ samples with probability at least 0.05 .

We start with a simplified version of our result, where we only have random access to the matrix entries. Below we extend this lower bound to the situation where we even have random access to the adjacency lists of all rows and columns.

Theorem 3.8. Let $C \geqslant 1$. Given an $n \times n$ binary matrix $A$ by random access to its entries, and given $\sqrt{\log (n) / n} \ll \phi \leqslant 1 / 100 C$ such that $\mathrm{OPT} /\|A\|_{0} \leqslant \phi$, computing a $(1+C \phi)$-approximation of OPT requires to read $\Omega(n / \phi)$ entries of $A$ (in the worst case over $A$ ).

Proof. Set $\phi^{\prime} \stackrel{\text { def }}{=} 25 C \phi$ and $k \stackrel{\text { def }}{=} \phi n / 2$. As in Lemma 3.7, consider binary random variables $X_{1}, \ldots, X_{k}$ with expectations $p_{1}, \ldots, p_{k}$, where $p_{i} \in\left\{1 / 2-\phi^{\prime}, 1 / 2+\phi^{\prime}\right\}$ for each $i$. We (implicitly) construct an $n \times n$ matrix $A$ as follows. For ever $k<i \leqslant n, 1 \leqslant j \leqslant n$ we set $A_{i, j} \stackrel{\text { def }}{=} 1$. For any $1 \leqslant i \leqslant k, 1 \leqslant j \leqslant n$ we sample a bit $b_{i, j}$ from $X_{i}$ and set $A_{i, j} \stackrel{\text { def }}{=} b_{i, j}$. Note that we can run any Binary $\ell_{0}$-Rank-1 algorithm implicitly on $A$ : whenever the algorithm reads an entry $A_{i, j}$ we sample a bit from $X_{i}$ to determine the entry (and we remember the entry for possible further accesses).

Let us determine the optimal solution for $A$. Note that for each $i>k$, since the row $A_{i, \text { : }}$ is all-ones, it is always better to pick this row than not to pick it, and thus without loss of generality any solution $u, v$ has $u_{i}=1$. Similarly, for any $j$, since the column $A_{:, j}$ has $n-k>n / 21$ 's in rows picked by $u$, it is always better to pick the column than not to pick it, and thus $v_{j}=1$, i.e., $v$ is the all-ones vector. Hence, the only choice is for any $1 \leqslant i \leqslant k$ to pick or not to pick row $i$. Note that no matter whether we pick these rows or not, the total error is at most $\phi n^{2} / 2$, since these rows in total have $k n=\phi n^{2} / 2$ entries, and all
remaining entries of $A$ are correctly recovered by the product $u v^{T}$ by the already chosen entries of $u$ and $v$. Hence, OPT $\leqslant \phi n^{2} / 2$, and since $\|A\|_{0} \geqslant(n-k) n \geqslant n^{2} / 2$, we obtain, as required, OPT/ $\|A\|_{0} \leqslant \phi$.

Now consider the rows $1 \leqslant i \leqslant k$ more closely. Since $v$ is the all-ones vector, not picking row $i$ incurs cost for each 1 in the row, which is cost $\left\|A_{i,:}\right\|_{0}$, while picking row $i$ incurs cost for each 0 in the row, which is cost $n-\left\|A_{i,:}\right\|_{0}$. Note that the expected number of 1 's in row $1 \leqslant i \leqslant k$ is $p_{i} n$. The Chernoff bound yields concentration: We have w.h.p. $\left|\left|A_{i,:} \|_{0}-p_{i} n\right| \leqslant 0.01 \cdot \phi^{\prime} n\right.$, where we used $\phi^{\prime} \gg \sqrt{\log (n) / n}$. In the following we condition on this event and thus drop "w.h.p." from our statements. In particular, for any $i$ with $p_{i}=1 / 2+\phi^{\prime}$ we have $\left\|A_{i,:}\right\|_{0} \geqslant\left(1 / 2+0.99 \phi^{\prime}\right) n$, and for any $i$ with $p_{i}=1 / 2-\phi^{\prime}$ we have $\left\|A_{i,:}\right\|_{0} \leqslant\left(1 / 2-0.99 \phi^{\prime}\right) n$.

By picking all rows $i \leqslant k$ with $p_{i}=1 / 2+\phi^{\prime}$ and not pick the rows with $p_{i}=1 / 2-\phi^{\prime}$, we see that $\mathrm{OPT} \leqslant\left(1 / 2-0.99 \phi^{\prime}\right) k n$. Now consider a solution $u$ that among the rows $1 \leqslant i \leqslant k$ with $p_{i}=1 / 2+\phi^{\prime}$ picks $g_{+}$many and does not pick $b_{+}$many. Similarly, among the rows with $p_{i}=1 / 2-\phi^{\prime}$ it picks $g_{-}$and does not pick $b_{-}$. Note that each of the $g_{+}$"good" rows incurs cost

$$
n-\left\|A_{i}:\right\|_{0} \geqslant n-\left(1 / 2+1.01 \phi^{\prime}\right) n=\left(1 / 2-1.01 \phi^{\prime}\right) n .
$$

Each of the $b_{+}$"bad" rows incurs a cost of $\left\|A_{i}:\right\|_{0} \geqslant\left(1 / 2+0.99 \phi^{\prime}\right) n$. Similar statements hold for $g_{-}$and $b_{-}$, and thus for $g \stackrel{\text { def }}{=} g_{+}+g_{-}$and $b \stackrel{\text { def }}{=} b_{+}+b_{-}$, with $g+b=k$, we obtain a total cost of

$$
\begin{aligned}
\left\|A-u v^{T}\right\|_{0} & \geqslant g \cdot\left(1 / 2-1.01 \phi^{\prime}\right) n+b \cdot\left(1 / 2+0.99 \phi^{\prime}\right) n \\
& =k\left(1 / 2-0.99 \phi^{\prime}\right) n+2 b \phi^{\prime} n-0.02 k \phi^{\prime} n \\
& \geqslant \mathrm{OPT}+2 b \phi^{\prime} n-0.02 \phi^{\prime} k n .
\end{aligned}
$$

If $b \geqslant 0.02 k$, then

$$
\left\|A-u v^{T}\right\|_{0} \geqslant \mathrm{OPT}+0.02 \phi^{\prime} k n \geqslant\left(1+0.04 \phi^{\prime}\right) \mathrm{OPT}
$$

By contraposition, if we compute a $\left(1+0.04 \phi^{\prime}=1+C \phi\right)$-approximation on $A$, then $b \leqslant 0.02 k$, and thus the vector $u$ correctly identifies for at least a 0.98 fraction of the random variables $X_{i}$ whether $p_{i}=1 / 2+\phi^{\prime}$ or $p_{i}=1 / 2-\phi^{\prime}$. Since this holds w.h.p., by Lemma 3.7 we need $\Omega\left(k / \phi^{\prime 2}\right)=\Omega\left(n /\left(\phi C^{2}\right)\right)$ samples from the variables $X_{i}$, and thus $\Omega\left(n /\left(\phi C^{2}\right)\right)$ reads in $A$. Since $C \geqslant 1$ is constant, we obtain a lower bound of $\Omega(n / \phi)$. This lower bound holds in expectation over the constructed distribution of $A$-matrices, and thus also in the worst case over $A$.

The construction of the above theorem does not work in case when we have random access to the adjacency lists of the rows, since this allows us to quickly determine the numbers of 1's per row, which is all we need to determine whether we want to pick a particular row in the matrix constructed above. To treat this issue, we adapt the construction as follows.

Proof of Theorem 1.7. We assume that $n$ is even. Let $\phi^{\prime}, k, X_{1}, \ldots, X_{k}, p_{1}, \ldots, p_{k}$ be as in the proof of Theorem 3.8. We adapt the construction of the matrix $A$ as follows. For any $2 k<i \leqslant n, 1 \leqslant j \leqslant n / 2$ we set $A_{i, 2 j} \stackrel{\text { def }}{=} 1$ and $A_{i, 2 j-1} \stackrel{\text { def }}{=} 0$. For any $1 \leqslant i \leqslant k, 1 \leqslant j \leqslant n / 2$ we sample a bit $b_{i, j}$ from $X_{i}$ and set $A_{2 i, 2 j} \stackrel{\text { def }}{=} A_{2 i-1,2 j-1} \stackrel{\text { def }}{=} b_{i, j}$ and $A_{2 i-1,2 j} \stackrel{\text { def }}{=} A_{2 i, 2 j-1} \stackrel{\text { def }}{=} 1-b_{i, j}$.

As before, when running any Binary $\ell_{0}$-Rank-1 algorithm on $A$ we can easily support random accesses to entries $A_{i, j}$, by sampling from $X_{\lceil i / 2\rceil}$ to determine the entry (and remembering the sampled bit for possible further accesses). Furthermore, we can now allow random accesses to the adjacency arrays of rows and columns. Specifically, if we want to determine the $\ell$-th 1 in row $i \leqslant 2 k$, we know that among the entries $A_{i, 1}, \ldots, A_{i, 2 \ell}$ there are exactly $\ell 1$ 's, since by construction $A_{i, 2 j-1}+A_{i, 2 j}=1$. Hence, the $\ell$-th 1 in row $i$ is at position $A_{i, 2 \ell-1}$ or $A_{i, 2 \ell}$, depending only on the sample $b_{\lceil i / 2\rceil, \ell}$ from $X_{\lceil i / 2\rceil}$. For rows $i>2 k$, the $\ell$-th 1 is simply at position $A_{i, 2 \ell}$. Thus, accessing the $\ell$-th 1 in any row takes at most one sample, so we can simulate any algorithm on $A$ with random access to the adjacency lists of rows. The situation for columns is essentially symmetric. Similarly, we can allow constant time access to the row and column sums.

In the remainder we show that the constructed matrix $A$ has essentially the same properties as the construction in Theorem 3.8. We first argue that any 2 -approximation $u, v$ for the Binary $\ell_{0}$-Rank- 1 problem on $A$ picks all rows $i>2 k$ and picks all even columns and does not pick any odd column. Thus, the only remaining choice is which rows $i \leqslant 2 k$ to pick. To prove this claim, first note that any solution following this pattern has error at most $2 k n=\phi n^{2}$, since the $2 k$ undecided rows have $2 k n$ entries, and all other entries are correctly recovered by the already chosen parts of $u v^{T}$. Hence, we have OPT $\leqslant \phi n^{2}$. Now consider any 2-approximation $u, v$, which must have cost at most $2 \phi n^{2}$. Note that $u$ picks at least
$(1-5 \phi) n$ of the rows $\{2 k+1, \ldots, n\}$, since each such row contains $n / 2$ 's that can only be recovered if we pick the row, so we can afford to ignore at most $8 k=4 \phi n$ of these $n-2 k=(1-\phi) n$ rows. Now, each even column contains at least $(1-5 \phi) n>n / 21$ 's in picked rows, and thus it is always better to pick the even columns. Similarly, each odd column contains at least $n / 20$ 's in picked rows, and thus it is always better not to pick the odd columns. Hence, we obtain without loss of generality $v_{2 j}=1$ and $v_{2 j-1}=0$. Finally, each row $i>2 k$ contains $n / 21$ 's in columns picked by $v$ and $n / 20$ 's in columns not picked by $v$, and thus it is always better to pick row $i$. Hence, we obtain without loss of generality $u_{i}=1$ for $i>2 k$.

Our goal now is to lower bound $\left\|A-u v^{T}\right\|_{0}$ in terms of OPT and the error term $b \phi^{\prime} n$, similarly to the proof in Theorem 3.8. Notice that we may ignore the odd columns, as they are not picked by $v$. Restricted to the even columns, row $2 i$ is exactly as row $i$ in the construction in Theorem 3.8, while row $2 i-1$ is row $2 i$ negated. Thus, analogously as in the proof of Theorem 3.8 , we obtain w.h.p. OPT $\leqslant\left(1 / 2-0.99 \phi^{\prime}\right) 2 k n$ and

$$
\left\|A-u v^{T}\right\|_{0} \geqslant \mathrm{OPT}+2 b \phi^{\prime} n-0.04 k \phi^{\prime} n \geqslant\left(1+0.04 \phi^{\prime}\right) \mathrm{OPT},
$$

where $b \geqslant 0.04 k$ is the number of "bad" rows $i \leqslant 2 k$. Again analogously, if we compute a $\left(1+0.04 \phi^{\prime}\right)=$ $1+C \phi)$-approximation on $A$, then $b \leqslant 0.04 k$, and thus w.h.p. we correctly identify for at least a 0.96 fraction of the random variables $X_{i}$ whether $p_{i}=1 / 2+\phi^{\prime}$ or $p_{i}=1 / 2-\phi^{\prime}$. As before, this yields a lower bound of $\Omega(n / \phi)$ samples.

### 3.3 Exact Algorithm

A variant of the algorithm from Theorem 1.4 can also be used to solve the Binary $\ell_{0}$-Rank- 1 problem exactly. This yields the following theorem, which in particular shows that the problem is in polynomial time when OPT $\leqslant O\left(\sqrt{\|A\|_{0}} \log (m n)\right)$.

Theorem 1.8 (from page 7). Given a matrix $A \in\{0,1\}^{m \times n}$, if $\mathrm{OPT} /\|A\|_{0} \leqslant 1 / 240$ then we can solve exactly the Binary $\ell_{0}-$ Rank-1 problem in time $2^{O\left(\mathrm{OPT} / \sqrt{\|A\|_{0}}\right)} \cdot \operatorname{poly}(m n)$.

Proof. This algorithm builds upon the algorithmic results established in Theorem 1.6 and Theorem 1.12, and it consists of the following three phases:

1. Run the algorithm in Theorem 1.12 to compute a 3 -approximation of $\psi=\mathrm{OPT} /\|A\|_{0}$, i.e. a number $\phi \in[\psi, 3 \psi]$.
2. Run Steps 1-4 of Algorithm 5, resulting in selected rows $R^{S}$ and columns $C^{S}$, and undecided rows $R^{\prime}=R^{R} \backslash R^{S}$ and columns $C^{\prime}=C^{R} \backslash C^{S}$. As shown above, the choices made by these steps are optimal.
3. For the remaining rows $R^{\prime}$ and columns $C^{\prime}$ we use brute force to find the optimum solution. Specifically, assume without loss of generality that $\left|R^{\prime}\right| \leqslant\left|C^{\prime}\right|$. Enumerate all binary vectors $u^{\prime} \in\{0,1\}^{R^{\prime}}$. For each $u^{\prime}$, set $\widetilde{u}_{i}=u_{i}^{\prime}$ for all $i \in R^{\prime}$ to complete the specification of a vector $\widetilde{u} \in\{0,1\}^{m}$. We can now find the optimal choice of vector $\widetilde{v}$ in polynomial time, since the optimal choice is to set $\widetilde{v}_{j}=1$ iff column $A_{:, j}$ has more 1 's than 0 's in the support of $\widetilde{u}$. Since some $u^{\prime}$ gives rise to the optimal vector $\widetilde{u}=u$, we solve the Binary $\ell_{0}$-Rank- 1 problem exactly.

To analyze the running time, note that by Claim 3.6 we have

$$
\min \left\{\left|R^{\prime}\right|,\left|C^{\prime}\right|\right\} \leqslant 6 \phi \min \{\alpha, \beta\} \leqslant 6 \phi \sqrt{\alpha \beta} .
$$

By Lemma 3.1.7 and $\phi \leqslant 3 \psi$, we obtain $\min \left\{\left|R^{\prime}\right|,\left|C^{\prime}\right|\right\}=O\left(\psi \sqrt{\|A\|_{0}}\right)$. Hence, we enumerate $2^{O\left(\psi \sqrt{\|A\|_{0}}\right)}=$ $2^{O\left(\mathrm{OPT} / \sqrt{\|A\|_{0}}\right)}$ vectors $u^{\prime}$, and the total running time is $2^{O\left(\mathrm{OPT} / \sqrt{\|A\|_{0}}\right)} \cdot \operatorname{poly}(m n)$. This completes the proof of Theorem 1.8.

## Algorithms For Reals $\ell_{0}$-Rank- $k$

Given a matrix $A \in \mathbb{R}^{m \times n}$ with $m \geqslant n$, an integer $k$ and an error $\varepsilon \in(0,1 / 2)$, our goal is to find an approximate solution $\widetilde{u} \in \mathbb{R}^{m}, \widetilde{v} \in \mathbb{R}^{n}$ of the Reals $\ell_{0}$-Rank- $k$ problem such that $\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0} \leqslant$ $(1+\varepsilon) \mathrm{OPT}_{k}$, where the optimal value is defined by

$$
\begin{equation*}
\mathrm{OPT}_{k} \stackrel{\text { def }}{=} \min _{U \in \mathbb{R}^{m \times k}, V \in \mathbb{R}^{k \times n}}\|A-U \cdot V\|_{0} . \tag{4.1}
\end{equation*}
$$

Organization In Section 4.1, we design a $\operatorname{poly}(k, \log n)$ bicriteria algorithm for the Reals $\ell_{0}$-Rank$k$ problem, which runs in polynomial time. In Section 4.2, we give an efficient $(2+\varepsilon)$-approximation algorithm for the Reals $\ell_{0}$-Rank- 1 problem.

### 4.1 Polytime Bicriteria Algorithm For Reals $\ell_{0}$-Rank- $k$

This section is organized as follows. In Subsection 4.1.1, we prove a structural lemma that guarantees the existence of $k$ columns that yield a $(k+1)$-approximation of $\mathrm{OPT}_{k}$, and we also give an $\Omega(k)$-approximation lower bound for any algorithm that selects $k$ columns from the input matrix $A$. In Subsection 4.1.2, we give an approximation algorithm that runs in time poly $\left(n^{k}, m\right)$ and achieves $O\left(k^{2}\right)$-approximation. To the best of our knowledge, this is the first algorithm with provable non-trivial approximation guarantees. In Subsection 4.1.3, we design a practical algorithm that runs in time $\operatorname{poly}(m, n)$ with an exponent independent of $k$, if we allow for a bicriteria solution.

### 4.1.1 Structural Results

We give a novel structural result showing that any matrix $A$ contains $k$ columns which provide a $(k+1)$ approximation for the Reals $\ell_{0}$-Rank- $k$ problem (4.1).

Lemma 4.1. Let $A \in \mathbb{R}^{m \times n}$ be a matrix and $k$ be an integer. There is a subset $J^{(k)} \subset[n]$ of size $k$ and a matrix $Z \in \mathbb{R}^{k \times n}$ such that $\left\|A-A_{:, J^{(k)}} \cdot Z\right\|_{0} \leqslant(k+1) \mathrm{OPT}_{k}$.

Proof. Let $Q^{(0)}$ be the set of columns $j$ with $U V_{:, j}=0$, and let $R^{(0)} \stackrel{\text { def }}{=}[n] \backslash Q^{(0)}$. Let $S^{(0)} \stackrel{\text { def }}{=}[n], T^{(0)} \stackrel{\text { def }}{=} \emptyset$. We split the value $\mathrm{OPT}_{k}$ into $\operatorname{OPT}\left(S^{(0)}, R^{(0)}\right) \stackrel{\text { def }}{=}\left\|A_{S^{(0)}, R^{(0)}}-U V_{S^{(0)}, R^{(0)}}\right\|_{0}$ and

$$
\operatorname{OPT}\left(S^{(0)}, Q^{(0)}\right) \stackrel{\text { def }}{=}\left\|A_{S^{(0)}, Q^{(0)}}-U V_{S^{(0)}, Q^{(0)}}\right\|_{0}=\left\|A_{S^{(0)}, Q^{(0)}}\right\|_{0}
$$

Suppose $\operatorname{OPT}\left(S^{(0)}, R^{(0)}\right) \geqslant\left|S^{(0)}\right|\left|R^{(0)}\right| /(k+1)$. Then, for any subset $J^{(k)}$ it follows that

$$
\min _{Z}\left\|A-A_{S^{(0)}, J^{(k)}} Z\right\|_{0} \leqslant\left|S^{(0)}\right|\left|R^{(0)}\right|+\left\|A_{S^{(0)}, Q^{(0)}}\right\|_{0} \leqslant(k+1) \mathrm{OPT}_{k}
$$

Otherwise, there is a column $i^{(1)}$ such that

$$
\left\|A_{S^{(0)}, i^{(1)}}-(U V)_{S^{(0)}, i^{(1)}}\right\|_{0} \leqslant \operatorname{OPT}\left(S^{(0)}, R^{(0)}\right) /\left|R^{(0)}\right| \leqslant \mathrm{OPT}_{k} /\left|R^{(0)}\right|
$$

Let $T^{(1)}$ be the set of indices on which $(U V)_{S^{(0)}, i^{(1)}}$ and $A_{S^{(0)}, i^{(1)}}$ disagree, and similarly $S^{(1)} \stackrel{\text { def }}{=}$ $S^{(0)} \backslash T^{(1)}$ on which they agree. Then we have $\left|T^{(1)}\right| \leqslant \mathrm{OPT}_{k} /\left|R^{(0)}\right|$. Hence, in the submatrix $T^{(1)} \times R^{(0)}$ the total error is at most $\left|T^{(1)}\right| \cdot\left|R^{(0)}\right| \leqslant \mathrm{OPT}_{k}$. Let $R^{(1)}, D^{(1)}$ be a partitioning of $R^{(0)}$ such that $A_{S^{(1)}, j}$ is linearly dependent on $A_{S^{(1)}, i^{(1)}}$ iff $j \in D^{(1)}$. Then by selecting column $A_{:, i^{(1)}}$ the incurred cost on matrix $S^{(1)} \times D^{(1)}$ is zero. For the remaining submatrix $S^{(\ell)} \times R^{(\ell)}$, we perform a recursive call of the algorithm.

We make at most $k$ recursive calls, on instances $S^{(\ell)} \times R^{(\ell)}$ for $\ell \in\{0, \ldots, k-1\}$. In the $\ell^{t h}$ iteration, either $\operatorname{OPT}\left(S^{(\ell)}, R^{(\ell)}\right) \geqslant\left|S^{(\ell)}\right|\left|R^{(\ell)}\right| /(k+1-\ell)$ and we are done, or there is a column $i^{(\ell+1)}$ which partitions $S^{(\ell)}$ into $S^{(\ell+1)}, T^{(\ell+1)}$ and $R^{(\ell)}$ into $R^{(\ell+1)}, D^{(\ell+1)}$ such that

$$
\left|S^{(\ell+1)}\right| \geqslant m \cdot \prod_{i=0}^{\ell}\left(1-\frac{1}{k+1-i}\right)=\frac{k-\ell}{k+1} \cdot m
$$

and for every $j \in D^{(\ell)}$ the column $A_{S^{(\ell+1), j}}$ belongs to the span of $\left\{A_{S^{(\ell+1), i^{(t)}}}\right\}_{t=1}^{\ell+1}$.
Suppose we performed $k$ recursive calls. We show now that the incurred cost in submatrix $S^{(k)} \times R^{(k)}$ is at most $\mathrm{OPT}\left(S^{(k)}, R^{(k)}\right) \leqslant \mathrm{OPT}_{k}$. By construction, $\left|S^{(k)}\right| \geqslant m /(k+1)$ and the sub-columns $\left\{A_{S^{(k), i}}\right\}_{i \in I^{(k)}}$ are linearly independent, where $I^{(k)}=\left\{i^{(1)}, \ldots, i^{(k)}\right\}$ is the set of the selected columns, and $A_{S^{(k)}, I^{(k)}}=$ $(U V)_{S^{(k)}, I^{(k)}}$. Since $\operatorname{rank}\left(A_{S^{(k)}, I^{(k)}}\right)=k$, it follows that $\operatorname{rank}\left(U_{S^{(k),:}}\right)=k, \operatorname{rank}\left(V_{:, I^{(k)}}\right)=k$ and the matrix $V_{:, I^{(k)}} \in \mathbb{R}^{k \times k}$ is invertible. Hence, for matrix $Z=\left(V_{:, I^{(k)}}\right)^{-1} V_{:, R^{k}}$ we have

$$
\operatorname{OPT}\left(S^{(k)}, R^{(k)}\right)=\left\|A_{S^{k}, R^{k}}-A_{S^{k}, I^{k}} Z\right\|_{0}
$$

The statement follows by noting that the recursive calls accumulate a total cost of at most $k \cdot \mathrm{OPT}_{k}$ in the submatrices $T^{(\ell+1)} \times R^{(\ell)}$ for $\ell \in\{0,1, \ldots, k-1\}$, as well as cost at most $\mathrm{OPT}_{k}$ in submatrix $S^{(k)} \times R^{(k)}$.

We now show that any algorithm that selects $k$ columns of a matrix $A$ incurs at least an $\Omega(k)$ approximation for the Reals $\ell_{0}$-Rank- $k$ problem.

Lemma 4.2. Let $k \leqslant n / 2$. Suppose $A=\left(G_{k \times n} ; I_{n \times n}\right) \in \mathbb{R}^{(n+k) \times n}$ is a matrix composed of a Gaussian random matrix $G \in \mathbb{R}^{k \times n}$ with $G_{i, j} \sim N(0,1)$ and identity matrix $I_{n \times n}$. Then for any subset $J^{(k)} \subset[n]$ of size $k$, we have $\min _{Z \in \mathbb{R}^{k \times n}}\left\|A-A_{:, J^{(k)}} \cdot Z\right\|_{0}=\Omega(k) \cdot \mathrm{OPT}_{k}$.
Proof. Notice that the optimum cost is at most $n$, achieved by selecting $U=\left(I_{k \times k} ; 0_{n \times k}\right)$ and $V=G_{k \times n}$. It is well known that Gaussian matrices are invertible with probability 1, see e.g. [SST06, Thm 3.3]. Hence, $G_{:, J^{(k)}}$ is a nonsingular matrix for every subset $J^{(k)} \subset[n]$ of size $k$.

We will show next that for any subset $J^{(k)}$ of $k$ columns the incurred cost is at least $(n-k) k \geqslant n k / 2$. Without loss of generality, the chosen columns $J^{(k)}=[k]$ are the first $k$ columns of $A$. Let $R=[2 k]$ be the first $2 k$ rows and $C=[n] \backslash J$ be the last $n-k$ columns. We bound

$$
\begin{aligned}
\min _{Z}\left\|A-A_{:,[k]} Z\right\|_{0} & \geqslant \min _{Z}\left\|A_{R, C}-A_{R,[k]} Z\right\|_{0} \\
& =\sum_{j \in C} \min _{z^{(j)}}\left\|A_{R, j}-A_{R,[k]} z^{(j)}\right\|_{0},
\end{aligned}
$$

i.e. we ignore all rows and columns except $R$ and $C$. Consider any column $j \in C$. Since $A_{R, j}=\left(G_{:, j}, 0_{k}\right)$ and $A_{R,[k]}=\left(G_{;,[k]}, I_{k \times k}\right)$, for any vector $z \in \mathbb{R}^{k}$ we have

$$
\begin{aligned}
\left\|A_{R, j}-A_{R, J^{(k)}} z\right\|_{0} & =\left\|G_{:, j}-G_{:,[k]} z\right\|_{0}+\left\|I_{k \times k} z\right\|_{0} \\
& =\left\|G_{:, j}-G_{:,[k]} z\right\|_{0}+\|z\|_{0} .
\end{aligned}
$$

Let $\ell \stackrel{\text { def }}{=}\|z\|_{0}$. By symmetry, without loss of generality we can assume that the first $\ell$ entries of $z$ are non-zero and the remaining entries are 0 . Let $x \in \mathbb{R}^{\ell}$ be the vector containing the first $\ell$ entries of $z$. Then we have

$$
\left\|A_{R, j}-A_{R, J^{(k)}} z\right\|_{0}=\left\|G_{:, j}-G_{:,[\ell]} x\right\|_{0}+\ell .
$$

We consider w.l.o.g. the first $k$ columns of $A$, and we construct the optimum matrix $Z$ that minimizes $\left\|A_{:, 1: k} Z-A\right\|_{0}$. Observe that it is optimal to set the first $k$ columns of $Z$ to $I_{k \times k}$, and since $A_{2 k+1: n, 1: k}=0$ we can focus only on the submatrix $A_{1: 2 k, k+1: n}=\left(G_{1: k, k+1: n} ; 0_{k \times n-k}\right)$.

Consider a column $A_{1: 2 k, j}$ for $j \in[k+1, n]$. Our goal is to find a vector $v \in \mathbb{R}^{k}$ minimizing the objective function $\Psi=\min _{v}\left\{\|v\|_{0}+\left\|G^{(k)} v-g\right\|_{0}\right\}$, where $G^{(k)} \stackrel{\text { def }}{=}\left\{G_{1: k, 1: k}\right\}$ and $g \xlongequal{=} G_{1: k, j}$. It holds with probability 1 that $G^{(k)}$ and $g$ do not have an entry equal to zero. Moreover, since $G^{(k)}$ is invertible every row in $\left[G^{(k)}\right]^{-1}$ is non-zero, and thus with probability 1 a vector $v=\left[G^{(k)}\right]^{-1} g$ has entry equal to zero.

Let $v=(x ; 0)$ be an arbitrary vector with $\|x\|_{1}=\ell$. Let $G^{(\ell)}$ be a submatrix of $G^{(k)}$ induced by the first $\ell$ columns. For every subset $S \subset[m]$ of $\ell$ rows the corresponding submatrix $G_{S,:}^{(\ell)}$ has a full rank.

Suppose there is a subset $S$ such that for $G_{S,:}^{(\ell)}$ and $g_{S}$ there is a vector $x \in \mathbb{R}^{k}$ satisfying $G_{S,:}^{(\ell)} x=g_{S}$. Since $G_{S,:}^{(\ell)}$ is invertible, the existence of $x$ implies its uniqueness. On the other hand, for any row $i \in[m] \backslash S$ the probability of the event $G_{i,:}^{(\ell)} x=g_{i}$ is equals to 0 . Since $G^{(k)} v=G^{(\ell)} x$ and there are finitely many possible subsets $S$ as above, i.e. $\binom{m}{\ell} \leqslant m^{\ell}$, by union bound it follows that $\left\|G^{(k)} v-g\right\|_{0} \geqslant k-\ell$. Therefore, it holds that $\phi \geqslant k$.

The statement follows by noting that the total cost incurred by $A_{;, 1: k}$ and any $Z$ is lower bounded by $(n-k) k+(n-k)=(1-k / n)(k+1) n$.

### 4.1.2 Basic Algorithm

We give an impractical algorithm that runs in time poly $\left(n^{k}, s m\right)$ and achieves $O\left(k^{2}\right)$-approximation. To the best of our knowledge this is the first approximation algorithm for the $\ell_{0}$-Rank- $k$ problem with non-trivial approximation guarantees.
Theorem 1.10 (from page 8). Given $A \in \mathbb{R}^{m \times n}$ and $k \in[n]$ we can compute in $O\left(n^{k+1} m^{2} k^{\omega+1}\right)$ time a set of $k$ indices $J^{(k)} \subset[n]$ and a matrix $Z \in \mathbb{R}^{k \times n}$ such that $\left\|A-A_{:, J(k)} \cdot Z\right\|_{0} \leqslant O\left(k^{2}\right) \cdot \mathrm{OPT}_{k}$.

We use as a subroutine the algorithm of Berman and Karpinski [BK02] (attributed also to Kannan in that paper) which given a matrix $U$ and a vector $b$ approximates $\min _{x}\|U x-b\|_{0}$ in polynomial time. Specifically, we invoke in our algorithm the following variant of this result established by Alon, Panigrahy, and Yekhanin [APY09].
Theorem 4.3. [APY09] There is an algorithm that given a matrix $A \in \mathbb{R}^{m \times k}$ and a vector $b \in \mathbb{R}^{m}$, outputs in time $O\left(m^{2} k^{\omega+1}\right)$ a vector $z \in \mathbb{R}^{k}$ such that w.h.p. $\|A z-b\|_{0} \leqslant k \cdot \min _{x}\|A x-b\|_{0}$.
Proof of Theorem 1.10. The existence of a subset $J^{*}$ of $k$ columns of $A$ and matrix $Z^{*} \in \mathbb{R}^{k \times n}$ with $\left\|A-A_{:, J^{*}} Z^{*}\right\|_{0} \leqslant(k+1) \mathrm{OPT}_{k}$ follows by Lemma 4.1. We enumerate all $\binom{n}{k}$ subsets $J^{(k)}$ of $k$ columns. For each $J^{(k)}$, we split $\min _{Z}\left\|A_{:, J^{(k)}} Z-A\right\|_{0}=\sum_{i=1}^{n} \min _{z^{(i)}}\left\|A_{:, J^{(k)}} z^{(i)}-A_{:, i}\right\|_{0}$, and we run the algorithm from Theorem 4.3 for each column $A_{:, i}$, obtaining approximate solutions $\widetilde{z}^{(1)}, \ldots, \widetilde{z}^{(n)}$ that form a matrix $\widetilde{Z}$. Then, we return the best solution $\left(A_{:, J^{(k)}}, \widetilde{Z}\right)$. To verify that this yields a $k(k+1)$-approximation, note that for $J^{(k)}=J^{*}$ we have

$$
\begin{aligned}
\left\|A_{:, J^{*}} \widetilde{Z}-A\right\|_{0} & =\sum_{i=1}^{n}\left\|A_{:, J^{*}} \widetilde{z}^{(i)}-A_{:, i}\right\|_{0} \leqslant k \sum_{i=1}^{n} \min _{z^{(i)}}\left\|A_{:, J^{*}} z^{(i)}-A_{:, i}\right\|_{0} \\
& =k \cdot \min _{Z}\left\|A_{:, J^{*}} Z-A\right\|_{0} \leqslant k(k+1) \cdot \mathrm{OPT}_{k}
\end{aligned}
$$

The time bound $O\left(n^{k+1} m^{2} k^{\omega+1}\right)$ is immediate from Theorem 4.3. This proves the statement.

### 4.1.3 Polynomial Time Bicriteria Algorithm

Our main contribution in this section is to design a practical algorithm that runs in time poly $(n, m)$ with an exponent independent of $k$, if we allow for a bicriteria solution.
Theorem 1.11 (from page 8). Given a matrix $A \in \mathbb{R}^{m \times n}$ and an integer $k$, there is an algorithm that in expected time $\operatorname{poly}(m, n)$ outputs a subset of indices $J \subset[n]$ with $|J|=O(k \log (n / k))$ and a matrix $Z \in \mathbb{R}^{|J| \times n}$ such that $\left\|A-A_{:, J} \cdot Z\right\|_{0} \leqslant O\left(k^{2} \log (n / k)\right) \cdot \mathrm{OPT}_{k}$.

The structure of the proof follows a recent approximation algorithm [CGK ${ }^{+} 17$, Algorithm 3] for the $\ell_{p}$-low rank approximation problem, for any $p \geqslant 1$. We note that the analysis of [CGK ${ }^{+}$17, Theorem 7] is missing an $O\left(\log ^{1 / p} n\right)$ approximation factor, and naïvely provides an $O\left(k \log ^{1 / p} n\right)$-approximation rather than the stated $O(k)$-approximation. Further, it might be possible to obtain an efficient algorithm yielding an $O\left(k^{2} \log k\right)$-approximation for Theorem 1.11 using unpublished techniques in [SWZ18]; we leave the study of obtaining the optimal approximation factor to future work.

There are two critical differences with the proof of $\left[\mathrm{CGK}^{+} 17\right.$, Theorem 7]. We cannot use the earlier $\left[\mathrm{CGK}^{+} 17\right.$, Theorem 3] which shows that any matrix $A$ contains $k$ columns which provide an $O(k)-$ approximation for the $\ell_{p}$-low rank approximation problem, since that proof requires $p \geqslant 1$ and critically uses scale-invariance, which does not hold for $p=0$. Our combinatorial argument in Lemma 4.1 seems fundamentally different than the maximum volume submatrix argument in [CGK $\left.{ }^{+} 17\right]$ for $p \geqslant 1$.

Second, unlike for $\ell_{p}$-regression for $p \geqslant 1$, the $\ell_{0}$-regression problem $\min _{x}\|U x-b\|_{0}$ given a matrix $U$ and vector $b$ is not efficiently solvable, since it corresponds to a nearest codeword problem which is

```
Algorithm 6 Bicriteria Algorithm: Selecting \(O(k \log (n / k))\) Columns
ApproximatelySelectColumns \((A, k)\)
    ensure \(A\) has at least \(k \log (n / k))\) columns
    . If the number of columns of matrix \(A\) is less than or equal to \(2 k\)
        Return all the columns of \(A\)
    Else
        Repeat
            Let \(R\) be a set of \(2 k\) uniformly random columns of \(A\)
        Until at least \(1 / 10\) fraction of columns of \(A\) are nearly approximately covered
        Let \(A_{\bar{R}}\) be the columns of \(A\) not nearly approximately covered by \(R\)
        Return \(R \cup\) ApproximatelySelectColumns \(\left(A_{\bar{R}}, k\right)\)
```

NP-hard [Ale11]. Thus, we resort to an approximation algorithm for $\ell_{0}$-regression, based on ideas for solving the nearest codeword problem in [APY09, BK02].

Note that $\mathrm{OPT}_{k} \leqslant\|A\|_{0}$. Since there are only $m n+1$ possibilities of $\mathrm{OPT}_{k}$, we can assume we know $\mathrm{OPT}_{k}$ and we can run the Algorithm 6 for each such possibility, obtaining a rank- $O(k \log n)$ solution, and then outputting the solution found with the smallest cost.

This can be further optimized by forming instead $O(\log (m n))$ guesses of $\mathrm{OPT}_{k}$. One of these guesses is within a factor of 2 from the true value of $\mathrm{OPT}_{k}$, and we note that the following argument only needs to know $\mathrm{OPT}_{k}$ up to a factor of 2 .

We start by defining the notion of approximate coverage, which is different than the corresponding notion in $\left[\mathrm{CGK}^{+} 17\right]$ for $p \geqslant 1$, due to the fact that $\ell_{0}$-regression cannot be efficiently solved. Consequently, approximate coverage for $p=0$ cannot be efficiently tested. Let $Q \subseteq[n]$ and $M=A_{:, Q}$ be an $m \times|Q|$ submatrix of $A$. We say that a column $M_{:, i}$ is $(S, Q)$-approximately covered by a submatrix $M_{:, S}$ of $M$, if $|S|=2 k$ and

$$
\begin{equation*}
\min _{x}\left\|M_{:, S} \cdot x-M_{:, i}\right\|_{0} \leqslant \frac{100(k+1) \mathrm{OPT}_{k}}{|Q|} . \tag{4.2}
\end{equation*}
$$

Lemma 4.4. (Similar to [CGK+17, Lemma 6], but using Lemma 4.1) Let $Q \subseteq[n]$ and $M=A_{:, Q}$ be a submatrix of $A$. Suppose we select a subset $R$ of $2 k$ uniformly random columns of $M$. Then with probability at least $1 / 3$, at least a $1 / 10$ fraction of the columns of $M$ are $(R, Q)$-approximately covered.

Proof. To show this, as in [CGK $\left.{ }^{+} 17\right]$, consider a uniformly random column index $i$ not in the set $R$. Let $T \stackrel{\text { def }}{=} R \cup\{i\}, \eta \stackrel{\text { def }}{=} \min _{\operatorname{rank}(B)=k}\left\|M_{:, T}-B\right\|_{0}$, and $B^{\star} \stackrel{\text { def }}{=} \arg \min _{\operatorname{rank}(B)=k}\|M-B\|_{0}$. Since $T$ is a uniformly random subset of $2 k+1$ columns of $M$, we have

$$
\begin{aligned}
\mathbb{E}_{T} \eta & \leqslant \mathbb{E}_{T}\left\|M_{:, T}-B_{:, T}^{\star}\right\|_{0}=\sum_{T \in\binom{|Q|}{2 k+1}} \sum_{i \in T}\left\|M_{:, i}-B_{:, i}^{\star}\right\|_{0} \operatorname{Pr}[T] \\
& =\sum_{i \in Q} \frac{\binom{|Q|-1}{|T-1|}}{\binom{|Q|}{|T|}}\left\|M_{:, i}-B_{:, i}^{\star}\right\|_{0}=\frac{(2 k+1) \mathrm{OPT}_{k}^{M}}{|Q|} \leqslant \frac{(2 k+1) \mathrm{OPT}_{k}}{|Q|} .
\end{aligned}
$$

Then, by a Markov bound, we have $\operatorname{Pr}\left[\eta \leqslant \frac{10(2 k+1) \mathrm{OPT}_{k}}{|Q|}\right] \geqslant 9 / 10$. Let $\mathcal{E}_{1}$ denotes this event.
Fix a configuration $T=R \cup\{i\}$ and let $L(T) \subset T$ be the subset guaranteed by Lemma 4.1 such that $|L(T)|=k$ and

$$
\min _{X}\left\|M_{:, L(T)} X-M_{:, T}\right\|_{0} \leqslant(k+1) \min _{\operatorname{rank}(B)=k}\left\|M_{:, T}-B\right\|_{0} .
$$

Notice that

$$
\mathbb{E}_{i}\left[\min _{x}\left\|M_{:, L(T)} x-M_{:, i}\right\|_{0} \mid T\right]=\frac{1}{2 k+1} \min _{X}\left\|M_{:, L(T)} X-M_{:, T}\right\|_{0},
$$

and thus by the law of total probability we have

$$
\mathbb{E}_{T}\left[\min _{x}\left\|M_{:, L(T)} x-M_{:, i}\right\|_{0}\right] \leqslant \frac{(k+1) \eta}{2 k+1}
$$

Let $\mathcal{E}_{2}$ denote the event that $\min _{x}\left\|M_{:, L} x-M_{:, i}\right\|_{0} \leqslant \frac{10(k+1) \eta}{2 k+1}$. By a Markov bound, $\operatorname{Pr}\left[\mathcal{E}_{2}\right] \geqslant 9 / 10$.

Further, as in $\left[\mathrm{CGK}^{+} 17\right]$, let $\mathcal{E}_{3}$ be the event that $i \notin L$. Observe that there are $\binom{k+1}{k}$ ways to choose a subset $R^{\prime} \subset T$ such that $\left|R^{\prime}\right|=2 k$ and $L \subset R^{\prime}$. Since there are $\binom{2 k+1}{2 k}$ ways to choose $R^{\prime}$, it follows that $\operatorname{Pr}[L \subset R \mid T]=\frac{k+1}{2 k+1}>1 / 2$. Hence, by the law of total probability, we have $\operatorname{Pr}\left[\mathcal{E}_{3}\right]>1 / 2$.

As in $\left[\mathrm{CGK}^{+} 17\right], \operatorname{Pr}\left[\mathcal{E}_{1} \wedge \mathcal{E}_{2} \wedge \mathcal{E}_{3}\right]>2 / 5$, and conditioned on $\mathcal{E}_{1} \wedge \mathcal{E}_{2} \wedge \mathcal{E}_{3}$,

$$
\begin{equation*}
\min _{x}\left\|M_{:, R} x-M_{:, i}\right\|_{0} \leqslant \min _{x}\left\|M_{:, L} x-M_{:, i}\right\|_{0} \leqslant \frac{10(k+1) \eta}{2 k+1} \leqslant \frac{100(k+1) \mathrm{OPT}_{k}}{|Q|}, \tag{4.3}
\end{equation*}
$$

where the first inequality uses that $L$ is a subset of $R$ given $\mathcal{E}_{3}$, and so the regression cost cannot decrease, while the second inequality uses the occurrence of $\mathcal{E}_{2}$ and the final inequality uses the occurrence of $\mathcal{E}_{1}$.

As in $\left[\mathrm{CGK}^{+} 17\right]$, if $Z_{i}$ is an indicator random variable indicating whether $i$ is approximately covered by $R$, and $Z=\sum_{i \in Q} Z_{i}$, then $\mathbb{E}_{R}[Z] \geqslant \frac{2|Q|}{5}$ and $\mathbb{E}_{R}[|Q|-Z] \leqslant \frac{3|Q|}{5}$. By a Markov bound, it follows that $\operatorname{Pr}\left[|Q|-Z \geqslant \frac{9|Q|}{10}\right] \leqslant \frac{2}{3}$. Thus, probability at least $1 / 3$, at least a $1 / 10$ fraction of the columns of $M$ are ( $R, Q$ )-approximately covered.

Given Lemma 4.4, we are ready to prove Theorem 1.11. As noted above, a key difference with the corresponding $\left[\mathrm{CGK}^{+} 17\right.$, Algorithm 3] for $\ell_{p}$ and $p \geqslant 1$, is that we cannot efficiently test if the $i$-th column is approximately covered by the set $R$. We will instead again make use of Theorem 4.3.

Proof of Theorem 1.11. The computation of matrix $Z$ force us to relax the notion of ( $R, Q$ )-approximately covered to the notion of $(R, Q)$-nearly-approximately covered as follows: we say that a column $M_{:, i}$ is ( $R, Q$ )-nearly-approximately covered if, the algorithm in Theorem 4.3 returns a vector $z$ such that

$$
\begin{equation*}
\left\|M_{:, R} z-M_{:, i}\right\|_{0} \leqslant \frac{100(k+1)^{2} \mathrm{OPT}_{k}}{|Q|} . \tag{4.4}
\end{equation*}
$$

By the guarantee of Theorem 4.3, if $M_{:, i}$ is $(R, Q)$-approximately covered then it is also with probability at least $1-1 / \operatorname{poly}(m n)(R, Q)$-nearly-approximately covered.

Suppose Algorithm 6 makes $t$ iterations and let $A_{:, \cup_{i=1}^{t} R_{i}}$ and $Z$ be the resulting solution. We bound now its cost. Let $B_{0}=[n]$, and consider the $i$-th iteration of Algorithm 6. We denote by $R_{i}$ a set of $2 k$ uniformly random columns of $B_{i-1}$, by $G_{i}$ a set of columns that is ( $R_{i}, B_{i-1}$ )-nearly-approximately covered, and by $B_{i}=B_{i-1} \backslash\left\{G_{i} \cup R_{i}\right\}$ a set of the remaining columns. By construction, $\left|G_{i}\right| \geqslant\left|B_{i-1}\right| / 10$ and

$$
\left|B_{i}\right| \leqslant \frac{9}{10}\left|B_{i-1}\right|-2 k<\frac{9}{10}\left|B_{i-1}\right| .
$$

Since Algorithm 6 terminates when $B_{t+1} \leqslant 2 k$, we have

$$
2 k<\left|B_{t}\right|<\left(1-\frac{1}{10}\right)^{t}\left|B_{0}\right|=\left(1-\frac{1}{10}\right)^{t} n
$$

and thus the number of iterations $t \leqslant 10 \log (n / 2 k)$. By construction, $\left|G_{i}\right|=\left(1-\alpha_{i}\right)\left|B_{i-1}\right|$ for some $\alpha_{i} \leqslant 9 / 10$, and hence

$$
\begin{equation*}
\sum_{i=1}^{t} \frac{\left|G_{i}\right|}{\left|B_{i-1}\right|} \leqslant t \leqslant 10 \log \frac{n}{2 k} \tag{4.5}
\end{equation*}
$$

Therefore, the solution cost is bounded by

$$
\begin{array}{cl} 
& \left\|A_{:, \cup_{i=1}^{t} R_{i}} Z-A\right\|_{0}=\sum_{i=1}^{t} \sum_{j \in G_{i}}\left\|A_{:, R_{i}} z^{(j)}-A_{:, j}\right\|_{0} \\
\stackrel{\text { Lem.4.4 }}{\leqslant} & \sum_{i=1}^{t} \sum_{j \in G_{i}} k \cdot \min _{x^{(j)}}\left\|A_{:, R_{i}} x^{(j)}-A_{:, j}\right\|_{0} \stackrel{(4.4)}{\lessgtr} \sum_{i=1}^{t} \sum_{j \in G_{i}} \frac{100(k+1)^{2} \mathrm{OPT}_{k}}{\left|B_{i-1}\right|} \\
= & 100(k+1)^{2} \mathrm{OPT}_{k} \cdot \sum_{i=1}^{t} \frac{\left|G_{i}\right|}{\left|B_{i-1}\right|} \stackrel{(4.5)}{\leqslant} O\left(k^{2} \cdot \log \frac{n}{2 k}\right) \cdot \mathrm{OPT}_{k} .
\end{array}
$$

By Lemma 4.4, the expected number of iterations of selecting a set $R_{i}$ such that $\left|G_{i}\right| \geqslant 1 / 10\left|B_{i-1}\right|$ is $O(1)$. Since the number of recursive calls $t$ is bounded by $O(\log (n / k))$, it follows by a Markov bound that Algorithm 6 chooses $O(k \log (n / k))$ columns in total. Since the approximation algorithm of Theorem 4.3 runs in polynomial time, our entire algorithm has expected polynomial time.

### 4.2 Approximation Algorithm for Reals $\ell_{0}$-Rank-1

Given a matrix $A \in \mathbb{R}^{m \times n}$ with $m \geqslant n$, and an error $\varepsilon \in(0,1 / 2)$, our goal is to find an approximate solution $\widetilde{u} \in \mathbb{R}^{m}, \widetilde{v} \in \mathbb{R}^{n}$ of the Reals $\ell_{0}$-Rank-1 problem such that $\left\|A-\widetilde{u} \cdot \widetilde{v}^{T}\right\|_{0} \leqslant(2+\varepsilon) \mathrm{OPT}_{1}$, where the optimal value is defined by

$$
\begin{equation*}
\mathrm{OPT}_{1} \stackrel{\text { def }}{=} \min _{u \in \mathbb{R}^{m}, v \in \mathbb{R}^{n}}\left\|A-u \cdot v^{T}\right\|_{0} \tag{4.6}
\end{equation*}
$$

In the trivial case when $\mathrm{OPT}_{1}=0$, there is an optimal algorithm that runs in time $O\left(\|A\|_{0}\right)$ and finds the exact rank-1 decomposition $u v^{T}$ of a matrix $A$. Here, we focus on the case when $\mathrm{OPT}_{1} \geqslant 1$. We will show that Algorithm 7 yields a $(2+\varepsilon)$-approximation and runs in nearly linear time in $\|A\|_{0}$, for any constant $\varepsilon>0$. Further, a variant of our algorithm even runs in sublinear time, if $\|A\|_{0}$ is large and

$$
\begin{equation*}
\psi \stackrel{\text { def }}{=} \mathrm{OPT}_{1} /\|A\|_{0} \tag{4.7}
\end{equation*}
$$

is not too small. In particular, we obtain sublinear time $o\left(\|A\|_{0}\right)$ when $\mathrm{OPT}_{1} \geqslant\left(\varepsilon^{-1} \log (m n)\right)^{4}$ and $\|A\|_{0} \geqslant n\left(\varepsilon^{-1} \log (m n)\right)^{4}$.
Theorem 1.12 (from page 9). There is an algorithm that, given $A \in \mathbb{R}^{m \times n}$ with column adjacency arrays and $\mathrm{OPT}_{1} \geqslant 1$, and given $\varepsilon \in(0,0.1]$, runs w.h.p. in time

$$
O\left(\left(\frac{n \log m}{\varepsilon^{2}}+\min \left\{\|A\|_{0}, n+\psi^{-1} \frac{\log n}{\varepsilon^{2}}\right\}\right) \frac{\log ^{2} n}{\varepsilon^{2}}\right)
$$

and outputs a column $A_{:, j}$ and a vector $z \in \mathbb{R}^{n}$ such that w.h.p. $\left\|A-A_{:, j} \cdot z^{T}\right\|_{0} \leqslant(2+\varepsilon) \mathrm{OPT}_{1}$. The algorithm also computes an estimate $Y$ satisfying w.h.p. $(1-\varepsilon) \mathrm{OPT}_{1} \leqslant Y \leqslant(2+2 \varepsilon) \mathrm{OPT}_{1}$.

In fact, our analysis of Theorem 1.12 directly applies to the Binary $\ell_{0}$-Rank- 1 problem, and yields as a special case the following result (which is used to prove Theorem 1.6 in Section 1.2).
Theorem 1.5 (from page 7). Let OPT $=\min _{u \in\{0,1\}^{m}, v \in\{0,1\}^{n}}\left\|A-u \cdot v^{T}\right\|_{0}$. Given a binary matrix $A \in\{0,1\}^{m \times n}$ with column adjacency arrays and $\mathrm{OPT} \geqslant 1$, and given $\varepsilon \in(0,0.1]$, we can compute w.h.p. in time

$$
O\left(\left(\frac{n \log m}{\varepsilon^{2}}+\min \left\{\|A\|_{0}, n+\psi^{-1} \frac{\log n}{\varepsilon^{2}}\right\}\right) \frac{\log ^{2} n}{\varepsilon^{2}}\right)
$$

a column $A_{:, j}$ and a binary vector $z \in\{0,1\}^{n}$ such that w.h.p. $\left\|A-A_{:, j} \cdot z^{T}\right\|_{0} \leqslant(2+\varepsilon)$ OPT. Further, we can compute an estimate $Y$ such that w.h.p. $(1-\varepsilon) \mathrm{OPT} \leqslant Y \leqslant(2+2 \varepsilon) \mathrm{OPT}$.

The rest of this section is devoted to proving Theorem 1.12. We start by presenting the pseudocode of Algorithm 7.

```
Algorithm 7 Reals \(\ell_{0}\)-Rank-1: Approximation Scheme
Input: \(A \in \mathbb{R}^{m \times n}\) and \(\varepsilon \in(0,0.1)\).
```

1. Partition the columns of $A$ into weight-classes $\mathcal{S}=\left\{S^{(0)}, \ldots, S^{(1+\log n)}\right\}$ such that
i) $S^{(0)}$ contains all columns $j$ with $\left\|A_{:, j}\right\|_{0}=0$, and
ii) $S^{(i)}$ contains all columns $j$ with $2^{i-1} \leqslant\left\|A_{:, j}\right\|_{0}<2^{i}$.
2. For each weight-class $S^{(i)}$ do:
2.1 Sample a set $C^{(i)}$ of $\Theta\left(\varepsilon^{-2} \log n\right)$ elements uniformly at random from $S^{(i)}$.
2.2 Find a $\left(1+\frac{\varepsilon}{15}\right)$-approximate solution $z^{(j)} \in \mathbb{R}^{n}$ for each column $A_{:, j} \in C^{(i)}$, i.e.

$$
\begin{equation*}
\left\|A-A_{:, j} \cdot\left[z^{(j)}\right]^{T}\right\|_{0} \leqslant\left(1+\frac{\varepsilon}{15}\right) \min _{v}\left\|A-A_{:, j} \cdot v^{T}\right\|_{0} \tag{4.8}
\end{equation*}
$$

3. Compute a $\left(1+\frac{\varepsilon}{15}\right)$-approximation $Y_{j}$ of $\left\|A-A_{:, j} \cdot\left[z^{(j)}\right]^{T}\right\|_{0}$ for every $j \in \bigcup_{i \in[|\mathcal{S}|]} C^{(i)}$.
4. Return the pair $\left(A_{:, j}, z^{(j)}\right)$ corresponding to the minimal value $Y_{j}$.

The only steps for which the implementation details are not immediate are Steps 2.2 and 3 . We will discuss them in Sections 4.2.2 and 4.2.3, respectively.

Note that the algorithm from Theorem 1.12 selects a column $A_{:, j}$ and then finds a good vector $z$ such that the product $A_{:, j} \cdot z^{T}$ approximates $A$. We show that the $(2+\varepsilon)$-approximation guarantee is essentially tight for algorithms following this pattern.

Lemma 4.5. There exist a matrix $A \in \mathbb{R}^{n \times n}$ such that $\min _{z}\left\|A-A_{:, j} \cdot z^{T}\right\|_{0} \geqslant 2(1-1 / n)$ OPT $_{1}$, for every column $A_{:, j}$.

Proof of Lemma 4.5. Let $A=I+J \in \mathbb{R}^{n \times n}$, where $I$ is an identity matrix and $J=\mathbf{1 1}^{T}$ is an allones matrix. Note that $\mathrm{OPT}_{1} \leqslant n$ is achieved by approximating $A$ with the rank- 1 matrix $J$. On the other hand, when we choose $u=A_{:, i}$ for any $i \in[n]$, the incurred cost on any column $A_{:, j}, j \neq i$, is $\min _{x}\left\|A_{:, j}-x A_{:, i}\right\|_{0}=2$, since there are two entries where $A_{:, i}$ and $A_{:, j}$ disagree. Hence, the total cost is at least $2 n-2 \geqslant(2-2 / n) \mathrm{OPT}_{1}$.

### 4.2.1 Correctness

We first prove the following structural result, capturing Steps 1-2.2 of Algorithm 7.
Lemma 4.6. Let $C^{(0)}, \ldots, C^{(\log n+1)}$ be the sets constructed in Step 2.1 of Algorithm 7, and let $C=$ $C^{(0)} \cup \ldots \cup C^{(\log n+1)}$. Then w.h.p. $C$ contains an index $j$ such that

$$
\min _{z}\left\|A-A_{:, j} \cdot z^{T}\right\|_{0} \leqslant(2+\varepsilon / 2) \mathrm{OPT}_{1}
$$

Proof. Let $u, v$ be an optimum solution of (4.6). For the weight class $S^{(0)}$ containing all columns without nonzero entries, setting $z_{c}=0$ for any $c \in S^{(0)}$ gives zero cost on these columns, no matter what column $A_{:, j}$ we picked. Hence, without loss of generality in the following we assume that $S^{(0)}=\emptyset$.

For any $i \geqslant 1$, we partition the weight class $S^{(i)}$ into $N^{(i)}, Z^{(i)}$ such that $v_{i}=0$ for every $i \in Z^{(i)}$ and $v_{i} \neq 0$ for $i \in N^{(i)}$. We denote by $\mathcal{S}^{+}$the set of weight-classes $S^{(i)}$ with $\left|N^{(i)}\right| \geqslant \frac{1}{3}\left|S^{(i)}\right|$. Let $\mathcal{R}=\bigcup_{i \in \mathcal{S}^{+}} S^{(i)}$ and $\mathcal{W}=[n] \backslash \mathcal{R}$. We partition $\mathcal{R}=\mathcal{N} \cup \mathcal{Z}$ such that $v_{i}=0$ for every $i \in \mathcal{Z}$ and $v_{i} \neq 0$ for $i \in \mathcal{N}$. Further, using the three sets $\mathcal{N}, \mathcal{Z}$ and $\mathcal{W}$ we decompose $\mathrm{OPT}_{1}$ into

$$
\begin{aligned}
\mathrm{OPT}_{1} & =\mathrm{OPT}_{\mathcal{N}}+\mathrm{OPT}_{\mathcal{Z}}+\mathrm{OPT}_{\mathcal{W}} \\
& =\left\|A_{:, \mathcal{N}}-u \cdot v_{\mathcal{N}}^{T}\right\|_{0}+\left\|A_{:, \mathcal{Z}}\right\|_{0}+\left\|A_{:, \mathcal{W}}-u \cdot v_{\mathcal{W}}^{T}\right\|_{0} .
\end{aligned}
$$

The proof proceeds by case distinction:

The set $\mathcal{Z}$ : For any column $A_{:, j}$ of $A$, we have

$$
\begin{equation*}
\min _{z \mathcal{Z}}\left\|A_{:, \mathcal{Z}}-A_{:, j} \cdot z_{\mathcal{Z}}^{T}\right\|_{0} \leqslant\left\|A_{:, \mathcal{Z}}-A_{:, j} \cdot 0\right\|_{0}=\left\|A_{:, \mathcal{Z}}\right\|_{0}=\mathrm{OPT}_{\mathcal{Z}} \tag{4.9}
\end{equation*}
$$

The set $\mathcal{W}$ : Note that $\mathcal{W}$ consists of all weight classes $S^{(i)}$ with $\left|Z^{(i)}\right|>\frac{2}{3}\left|S^{(i)}\right|$. For any such weight class $S^{(i)}$, the optimum cost satisfies

$$
\left\|A_{:, S^{(i)}}-u v_{S^{(i)}}^{T}\right\|_{0} \geqslant\left\|A_{:, Z^{(i)}}\right\|_{0} \geqslant \frac{2}{3}\left|S^{(i)}\right| 2^{i-1}=\frac{1}{3}\left|S^{(i)}\right| 2^{i} .
$$

Further, for any column $A_{:, j}$ of $A$, we have

$$
\begin{aligned}
\min _{z}\left\|A_{:, S^{(i)}}-A_{:, j} z^{T}\right\|_{0} & \leqslant\left\|A_{:, S^{(i)}}\right\|_{0} \leqslant\left\|A_{:, Z^{(i)}}\right\|_{0}+\frac{1}{3}\left|S^{(i)}\right| 2^{i} \\
& \leqslant 2\left\|A_{:, Z^{(i)}}\right\|_{0} \leqslant 2\left\|A_{:, S^{(i)}}-u v_{S^{(i)}}^{T}\right\|_{0}
\end{aligned}
$$

and thus the total cost in $\mathcal{W}$ is bounded by

$$
\begin{equation*}
\min _{z \mathcal{W}}\left\|A_{:, \mathcal{W}}-A_{:, j} \cdot z_{\mathcal{W}}^{T}\right\|_{0}=\sum_{i \in \mathcal{W}} \min _{z_{S}(i)}\left\|A_{:, S^{(i)}}-A_{:, j} \cdot z_{S^{(i)}}^{T}\right\|_{0} \leqslant 2\left\|A_{:, \mathcal{W}}-u v_{\mathcal{W}}^{T}\right\|_{0}=2 \mathrm{OPT}_{\mathcal{W}} \tag{4.10}
\end{equation*}
$$

The set $\mathcal{N}$ : By an averaging argument there is a subset $G \subseteq \mathcal{N}$ of size $|G| \geqslant \frac{\varepsilon}{3}|\mathcal{N}|$ such that for every $j \in G$ we have

$$
\left\|A_{:, j}-v_{j} \cdot u\right\|_{0} \leqslant \frac{1}{1-\varepsilon / 3} \cdot \frac{\mathrm{OPT}_{\mathcal{N}}}{|\mathcal{N}|} \leqslant\left(1+\frac{\varepsilon}{2}\right) \cdot \frac{\mathrm{OPT}_{\mathcal{N}}}{|\mathcal{N}|}
$$

Let $j \in G$ be arbitrary. Furthermore, let $P^{(j)}$ be the set of all rows $i$ with $A_{i, j}=v_{j} \cdot u_{i}$, and let $Q^{(j)}=[m] \backslash P^{(j)}$. By construction, we have $\left|Q^{(j)}\right| \leqslant\left(1+\frac{\varepsilon}{2}\right)$ OPT $_{\mathcal{N}} /|\mathcal{N}|$. Moreover, since $j \in \mathcal{N}$ we have
$v_{j} \neq 0$, and thus we may choose $z_{\mathcal{N}}=\frac{1}{v_{j}} v_{\mathcal{N}}$. This yields

$$
\begin{align*}
\min _{z_{\mathcal{N}}}\left\|A_{:, \mathcal{N}}-A_{:, j} \cdot z_{\mathcal{\mathcal { N }}}^{T}\right\|_{0} & \leqslant\left\|A_{:, \mathcal{N}}-A_{:, j} \cdot \frac{1}{v_{j}} v_{\mathcal{N}}^{T}\right\|_{0} \\
& =\left\|A_{P^{(j)}, \mathcal{N}}-u_{P^{(j)}} \cdot v_{\mathcal{N}}^{T}\right\|_{0}+\left\|A_{Q^{(j)}, \mathcal{N}}-A_{:, j} \cdot \frac{1}{v_{j}} v_{\mathcal{N}}^{T}\right\|_{0} \\
& \leqslant \mathrm{OPT}_{\mathcal{N}}+\left(1+\frac{\varepsilon}{3}\right) \mathrm{OPT}_{\mathcal{N}}=\left(2+\frac{\varepsilon}{3}\right) \mathrm{OPT}_{\mathcal{N}} \tag{4.11}
\end{align*}
$$

Hence, by combining (4.9), (4.10) and (4.11), it follows for any index $j \in G$ that

$$
\begin{aligned}
& \min _{z}\left\|A-A_{:, j} \cdot z^{T}\right\|_{0} \\
= & \min _{z \mathcal{W}}\left\|A_{:, \mathcal{W}}-A_{:, j} \cdot z_{\mathcal{W}}^{T}\right\|_{0}+\min _{z \mathcal{Z}}\left\|A_{:, \mathcal{Z}}-A_{:, j} \cdot z_{\mathcal{Z}}^{T}\right\|_{0}+\min _{z_{\mathcal{N}}}\left\|A_{:, \mathcal{N}}-A_{:, j} \cdot z_{\mathcal{N}}^{T}\right\|_{0} \\
\leqslant & 2 \mathrm{OPT}_{\mathcal{W}}+\mathrm{OPT}_{\mathcal{Z}}+\left(2+\frac{\varepsilon}{2}\right) \mathrm{OPT}_{\mathcal{N}} \leqslant\left(2+\frac{\varepsilon}{2}\right) \mathrm{OPT}_{1}
\end{aligned}
$$

This yields the desired approximation guarantee, provided that we sampled a column $A_{:, j}$ from $G$. We show next that whenever $\mathcal{N} \neq \emptyset$, our algorithm samples with high probability at least one column from $G$.

Before that let us consider the case when $\mathcal{N}=\emptyset$. Then, since the bounds (4.9) and (4.10) hold for any column $A_{:, j}$ of $A$, the set $C$ contains only good columns. Thus, we may assume that $\mathcal{N} \neq \emptyset$.

We now analyze the probability of sampling a column $A_{i, j}$ from $G$. By construction, the set $\mathcal{N}$ is the union of all $N^{(i)}$ such that $\left|N^{(i)}\right| \geqslant \frac{1}{3}\left|S^{(i)}\right|$. As shown above, we have $|G| \geqslant \frac{\varepsilon}{3}|\mathcal{N}|$, and thus there is an index $i$ satisfying $\left|G \cap S^{(i)}\right| \geqslant \frac{\varepsilon}{3}\left|N^{(i)}\right| \geqslant \frac{\varepsilon}{9}\left|S^{(i)}\right|$. Hence, when sampling a uniformly random element from $S^{(i)}$ we hit $G$ with probability at least $\frac{\varepsilon}{9}$. Since we sample $\Theta\left(\varepsilon^{-2} \log n\right)$ elements from $S^{(i)}$, we hit $G$ with high probability. This finishes the proof.

Correctness Proof of Algorithm 7: It remains to show that the pair $\left(A_{:, j}, z^{j}\right)$ with minimum estimate $Y_{j}$ yields a $(2+\varepsilon)$-approximation to $\mathrm{OPT}_{1}$. By Step 3, for every column $j$ we have

$$
\begin{equation*}
\left(1+\frac{\varepsilon}{15}\right)^{-1} \cdot\left\|A-A_{:, j}\left[z^{(j)}\right]^{T}\right\|_{0} \leqslant Y_{j} \leqslant\left(1+\frac{\varepsilon}{15}\right) \cdot\left\|A-A_{:, j}\left[z^{(j)}\right]^{T}\right\|_{0} \tag{4.12}
\end{equation*}
$$

Since $Y_{j} \leqslant Y_{j^{\prime}}$ for any other column $j^{\prime}$, (4.12) and the approximation guarantee of Steps 2.2 yield

$$
\left(1+\frac{\varepsilon}{15}\right)^{-1}\left\|A-A_{:, j}\left[z^{(j)}\right]^{T}\right\|_{0} \leqslant\left(1+\frac{\varepsilon}{15}\right)\left\|A-A_{:, j^{\prime}}\left[z^{\left(j^{\prime}\right)}\right]^{T}\right\|_{0} \leqslant\left(1+\frac{\varepsilon}{15}\right)^{2} \min _{z}\left\|A-A_{:, j^{\prime}} z^{T}\right\|_{0}
$$

By Lemma 4.6, w.h.p. there exists a column $j^{\prime} \in C$ with $\min _{z}\left\|A-A_{:, j^{\prime}} z^{T}\right\|_{0} \leqslant\left(2+\frac{\varepsilon}{2}\right) \mathrm{OPT}_{1}$. We obtain a total approximation ratio of $\left(1+\frac{\varepsilon}{15}\right)^{3}\left(2+\frac{\varepsilon}{2}\right) \leqslant 2+\varepsilon$ for any error $0<\varepsilon \leqslant 0.1$, i.e. we have $\left\|A-A_{:, j}\left[z^{(j)}\right]^{T}\right\|_{0} \leqslant(2+\varepsilon) \mathrm{OPT}_{1}$. Therefore, it holds that

$$
(1-\varepsilon) \mathrm{OPT}_{1} \leqslant\left(1+\frac{\varepsilon}{15}\right)^{-1} \mathrm{OPT}_{1} \leqslant Y_{j} \leqslant\left(1+\frac{\varepsilon}{15}\right)(2+\varepsilon) \mathrm{OPT}_{1} \leqslant(2+2 \varepsilon) \mathrm{OPT}_{1}
$$

This finishes the correctness proof.

### 4.2.2 Implementing Step 2.2

Step 2.2 of Algorithm 7 uses the following sublinear procedure.

```
Algorithm 8 Reals \(\ell_{0}\)-Rank-1: Objective Value Estimation
Input: \(A \in \mathbb{R}^{m \times n}, u \in \mathbb{R}^{m}\) and \(\varepsilon \in(0,1)\).
    let \(t \stackrel{\text { def }}{=} \Theta\left(\varepsilon^{-2} \log m\right), N \stackrel{\text { def }}{=} \operatorname{supp}(u)\), and \(p \stackrel{\text { def }}{=} t /|N|\).
1. Select each index \(i \in N\) with probability \(p\) and let \(S\) be the resulting set.
2. Compute a vector \(z \in \mathbb{R}^{n}\) such that \(z_{j}=\arg \min _{r \in \mathbb{R}}\left\|A_{S, j}-r \cdot u_{S}\right\|_{0}\) for all \(j \in[n]\).
3. Return the vector \(z\).
```

We prove now the correctness of Algorithm 8 and we analyze its runtime.

Lemma 4.7. Given $A \in \mathbb{R}^{m \times n}, u \in \mathbb{R}^{m}$ and $\varepsilon \in(0,1)$ we can compute in time $O\left(\varepsilon^{-2} n \log m\right)$ a vector $z \in \mathbb{R}^{n}$ such that w.h.p. for every $i \in[n]$ it holds that

$$
\left\|A_{:, i}-z_{i} u\right\|_{0} \leqslant(1+\varepsilon) \min _{v_{i} \in \mathbb{R}}\left\|A_{:, i}-v_{i} u\right\|_{0}
$$

Proof. Let $N, Z$ be a partitioning of $[m]$ such that $u_{i}=0$ for $i \in Z$ and $u_{i} \neq 0$ for $i \in N$. Since $\left\|A-u \cdot z^{T}\right\|_{0}=\left\|A_{N,:}-u_{N} \cdot z^{T}\right\|_{0}+\left\|A_{Z,:}\right\|_{0}$, it suffices to find a vector $z$ such that for every $j \in[n]$ we have

$$
\begin{equation*}
\left\|A_{N, j}-z_{j} \cdot u_{N}\right\|_{0} \leqslant(1+\varepsilon) \cdot \min _{v_{j}}\left\|A_{N, j}-v_{j} \cdot u_{N}\right\|_{0} \tag{4.13}
\end{equation*}
$$

Let $j \in[n]$ be arbitrary. For $r \in \mathbb{R}$ let $G(r) \stackrel{\text { def }}{=}\left\{i \in N: A_{i, j} / u_{i}=r\right\}$ be the set of entries of $A_{N, j}$ that we correctly recover by setting $z_{j}=r$. Note that $\left\|A_{N, j}-z_{j} \cdot u_{N}\right\|_{0}=|N|-\left|G\left(z_{j}\right)\right|$ holds for any $z_{j} \in \mathbb{R}$. Hence, the optimal solution sets $z_{j}=r^{\star} \stackrel{\text { def }}{=} \arg \max _{r \in \mathbb{R}}|G(r)|$.

Let $X_{G(r)}$ be a random variable indicating the number of elements selected from group $G(r)$ in Step 1 of Algorithm 8. Notice that $\mathbb{E}\left[X_{G(r)}\right]=t \cdot|G(r)| /|N|$, and by Chernoff bound w.h.p. we have

$$
\begin{equation*}
\left|X_{G(r)}-\mathbb{E}\left[X_{G(r)}\right]\right| \leqslant(\varepsilon / 8) \cdot t \tag{4.14}
\end{equation*}
$$

Let $S \subseteq N$ be the set of selected indices. Further, since $|S|=\sum_{\ell} X_{G(r)}$ and $\mathbb{E}[|S|]=t$, by Chernoff bound we have w.h.p. $|S| \leqslant(1+\varepsilon)|t|$. Observe that Step 2 of Algorithm 8 selects $z_{j}=\arg \max _{r \in \mathbb{R}} X_{G(r)}$, since $\left\|A_{S, j}-r \cdot u_{S}\right\|_{0}=|S|-X_{G(r)}$. We now relate $z_{j}$ to $r^{\star}$. The proof proceeds by case distinction on $\delta^{\star} \stackrel{\text { def }}{=}\left|G\left(r^{\star}\right)\right| /|N|$.

Case 1: Suppose $\delta^{\star} \leqslant \varepsilon / 4$. Then $\left\|A_{N, j}-r \cdot u_{N}\right\|_{0} \geqslant(1-\varepsilon / 4)|N|$ for every $r \in \mathbb{R}$, and thus no matter which $z_{j}$ is selected we obtain a $(1+\varepsilon)$-approximation, since

$$
\left\|A_{N, j}-z_{j} \cdot u_{N}\right\|_{0} \leqslant|N| \leqslant(1+\varepsilon) \min _{r}\left\|A_{N, j}-r \cdot u_{N}\right\|_{0}
$$

Case 2: Suppose $\delta^{\star} \geqslant 1 / 2+\varepsilon$. Then, by (4.14) w.h.p. we have

$$
X_{G\left(r^{\star}\right)} \geqslant \mathbb{E}\left[X_{G\left(r^{\star}\right)}\right]-(\varepsilon / 8) t=\left(\delta^{\star}-\varepsilon / 8\right) t>(1+\varepsilon) t / 2 \geqslant|S| / 2
$$

and thus $X_{r^{\star}}$ is maximal among all $X_{r}$. Hence, we select the optimal $z_{j}=r^{\star}$.
Case 3: Suppose $\varepsilon / 4<\delta^{\star}<1 / 2+\varepsilon$. Let $z_{j}=r$ be the value chosen by Algorithm 8. By (4.14), the event of making a mistake, given by $X_{G(r)} \geqslant X_{G\left(r^{\star}\right)}$, happens when

$$
\begin{equation*}
\mathbb{E}\left[X_{G(r)}\right]+(\varepsilon / 8) t \geqslant \mathbb{E}\left[X_{G\left(r^{\star}\right)}\right]-(\varepsilon / 8) t \tag{4.15}
\end{equation*}
$$

Let $\delta \stackrel{\text { def }}{=}|G(r)| /|N|$ and note that (4.15) implies $\delta \geqslant \delta^{\star}-\varepsilon / 4$. Hence, for the selected $r \neq r^{\star}$ we have

$$
\begin{aligned}
\left\|A_{N, j}-r \cdot u_{N}\right\|_{0} & =(1-\delta)|N| \leqslant\left(1-\delta^{\star}+\varepsilon / 4\right)|N| \\
& \leqslant(1+\varepsilon)\left(1-\delta^{\star}\right)|N|=(1+\varepsilon)\left\|A_{N, j}-r^{\star} \cdot u_{N}\right\|_{0}
\end{aligned}
$$

Therefore, in each of the preceding three cases, we obtain w.h.p. a $(1+\varepsilon)$-approximate solution. The statement follows by the union bound.

### 4.2.3 Implementing Step 3

In Step 3 of Algorithm 7 we want to compute a $\left(1+\frac{\varepsilon}{15}\right)$-approximation $Y_{j}$ of $\left\|A-A_{:, j} \cdot\left[z^{(j)}\right]^{T}\right\|_{0}$ for every $j \in \bigcup_{i \in[|\mathcal{S}|]} C^{(i)}$. We present two solutions, an exact algorithm (see Lemma 4.8) and a sublinear time sampling-based algorithm (see Lemma 4.10).

Lemma 4.8. Suppose $A, B \in \mathbb{R}^{m \times n}$ are represented by column adjacency arrays. Then, we can compute in time $O\left(\|A\|_{0}+n\right)$ the measure $\|A-B\|_{0}$.
Proof. We partition the entries of $A$ into five sets:

$$
\begin{array}{ll}
T_{1}=\left\{(i, j): A_{i j}=0 \text { and } B_{i j} \neq 0\right\}, & T_{4}=\left\{(i, j): 0 \neq A_{i j}=B_{i j} \neq 0\right\}, \\
T_{2}=\left\{(i, j): A_{i j} \neq 0 \text { and } B_{i j}=0\right\}, & T_{5}=\left\{(i, j): A_{i j}=B_{i j}=0\right\} \\
T_{3}=\left\{(i, j): 0 \neq A_{i j} \neq B_{i j} \neq 0\right\} . &
\end{array}
$$

Observe that $\|A-B\|_{0}=\left|T_{1}\right|+\left|T_{2}\right|+\left|T_{3}\right|$ and $\|B\|_{0}=\left|T_{1}\right|+\left|T_{3}\right|+\left|T_{4}\right|$. Since $\|A-B\|_{0}=$ $\|B\|_{0}+\left|T_{2}\right|-\left|T_{4}\right|$, it suffices to compute the numbers $\|B\|_{0},\left|T_{2}\right|$ and $\left|T_{4}\right|$. We compute $\left|T_{2}\right|$ and $\left|T_{4}\right|$ in $O\left(\|A\|_{0}\right)$ time, by enumerating all non-zero entries of $A$ and performing $O(1)$ checks for each. For $\|B\|_{0}$, we sum the column lengths of $B$ in time $O(n)$.

For our second, sampling-based implementation of Step 3, we make use of an algorithm by Dagum et al. [DKLR00] for estimating the expected value of a random variable. We note that the runtime of their algorithm is a random variable, the magnitude of which is bounded w.h.p. within a certain range.

Theorem 4.9. [DKLR00] Let $X$ be a random variable taking values in $[0,1]$ with $\mu \stackrel{\text { def }}{=} \mathbb{E}[X]>0$. Let $0<\varepsilon, \delta<1$ and $\rho_{X}=\max \{\operatorname{Var}[X], \varepsilon \mu\}$. There is an algorithm with sample access to $X$ that computes an estimator $\widetilde{\mu}$ in time $t$ such that for a universal constant $c$ we have
i) $\operatorname{Pr}[(1-\varepsilon) \mu \leqslant \widetilde{\mu} \leqslant(1+\varepsilon) \mu] \geqslant 1-\delta$
and
ii) $\operatorname{Pr}\left[t \geqslant c \cdot \varepsilon^{-2} \mu^{-2} \rho_{X} \log (1 / \delta)\right] \leqslant \delta$.

We state now our key technical insight, on which we build upon our sublinear algorithm.
Lemma 4.10. There is an algorithm that, given $A, B \in \mathbb{R}^{m \times n}$ with column adjacency arrays and $\|A-B\|_{0} \geqslant 1$, and given $\varepsilon>0$, computes an estimator $Z$ that satisfies w.h.p.

$$
(1-\varepsilon)\|A-B\|_{0} \leqslant Z \leqslant(1+\varepsilon)\|A-B\|_{0} .
$$

The algorithm runs w.h.p. in time $\left.O\left(n+\varepsilon^{-2} \frac{\|A\|_{0}+\|B\|_{0}}{\|A-B\|_{0}} \log n\right\}\right)$.
Proof. By Lemma 1.9, after $O(n)$ preprocessing time we can sample a uniformly random non-zero entry from $A$ or $B$ in time $O(1)$.

We consider the following random process:

1. Sample $C \in\{A, B\}$ such that $\operatorname{Pr}[C=A]=\frac{\|A\|_{0}}{\|A\|_{0}+\|B\|_{0}}$ and $\operatorname{Pr}[C=B]=\frac{\|B\|_{0}}{\|A\|_{0}+\|B\|_{0}}$.
2. Sample ( $i, j$ ) uniformly at random from the non-zero entries of $C$
3. Return:

$$
X= \begin{cases}0, & \text { if } A_{i j}=B_{i j} \\ 1 / 2, & \text { if } 0 \neq A_{i j} \neq B_{i j} \neq 0 \\ 1, & \text { if } A_{i j} \neq B_{i j} \text { and either } A_{i j} \text { or } B_{i j} \text { equals } 0 .\end{cases}
$$

Observe that

$$
\begin{aligned}
\mathbb{E}[X]= & \sum_{(i, j): A_{i j} \neq B_{i j}=0} \frac{\|A\|_{0}}{\|A\|_{0}+\|B\|_{0}} \cdot \frac{1}{\|A\|_{0}}+\sum_{(i, j): 0=A_{i j} \neq B_{i j}} \frac{\|B\|_{0}}{\|A\|_{0}+\|B\|_{0}} \cdot \frac{1}{\|B\|_{0}} \\
& +\sum_{(i, j): 0 \neq A_{i j} \neq B_{i j} \neq 0}\left(\frac{1}{2} \cdot \frac{\|A\|_{0}}{\|A\|_{0}+\|B\|_{0}} \cdot \frac{1}{\|A\|_{0}}+\frac{1}{2} \cdot \frac{\|B\|_{0}}{\|A\|_{0}+\|B\|_{0}} \cdot \frac{1}{\|B\|_{0}}\right) \\
= & \frac{\|A-B\|_{0}}{\|A\|_{0}+\|B\|_{0}} .
\end{aligned}
$$

Straightforward checking shows that $X \in[0,1]$ implies $\operatorname{Var}[X] \leqslant \mathbb{E}[X]$, and thus

$$
\rho_{X}=\max \{\operatorname{Var}[X], \varepsilon \cdot \mathbb{E}[X]\} \leqslant \mathbb{E}[X]
$$

Theorem 4.9 applied with $\delta=1 / \operatorname{poly}(n)$, yields w.h.p. in $O\left(\varepsilon^{-2} \mathbb{E}[X]^{-1} \log n\right)=O\left(\varepsilon^{-2} \frac{\|A\|_{0}+\|B\|_{0}}{\|A-B\|_{0}} \log n\right)$ time an estimator $(1-\varepsilon) \mathbb{E}[X] \leqslant \widetilde{\mu} \leqslant(1+\varepsilon) \mathbb{E}[X]$. Then, w.h.p. the estimator $Z \stackrel{\text { def }}{=}\left(\|A\|_{0}+\|B\|_{0}\right) \widetilde{\mu}$ satisfies the statement.

We present now our main result in this section.
Theorem 4.11. There is an algorithm that, given $A \in \mathbb{R}^{m \times n}$ with column adjacency arrays and $\mathrm{OPT}_{1} \geqslant$ 1 , and given $j \in[n], v \in \mathbb{R}^{m}$ and $\varepsilon \in(0,1)$, outputs an estimator $Y$ that satisfies w.h.p.

$$
(1-\varepsilon)\left\|A-A_{:, j} \cdot v^{T}\right\|_{0} \leqslant Y \leqslant(1+\varepsilon)\left\|A-A_{:, j} \cdot v^{T}\right\|_{0}
$$

The algorithm runs w.h.p. in time $O\left(\min \left\{\|A\|_{0}, n+\varepsilon^{-2} \psi^{-1} \log n\right\}\right)$, where $\psi=\mathrm{OPT}_{1} /\|A\|_{0}$.

Proof. Let $B \stackrel{\text { def }}{=} A_{:, j} v^{T}$ and observe that $\|A-B\|_{0} \geqslant \mathrm{OPT}_{1} \geqslant 1$. Note that we implicitly have column adjacency arrays for $B$, since for any column $c$ with $v_{c}=0$ there are no non-zero entries in $B_{:, c}$, and for any column $c$ with $v_{c}=1$ the non-zero entries of $B_{:, c}$ are the same as for $A_{:, j}$. Hence, Lemma 4.8 and Lemma 4.10 are applicable.

We analyze the running time of Lemma 4.10. Note that if $\|B\|_{0} \leqslant(1+\psi)\|A\|_{0}$ then $\frac{\|A\|_{0}+\|B\|_{0}}{\|A-B\|_{0}} \leqslant$ $(2+\psi) / \psi$, and otherwise, i.e. $(1+\psi)\|A\|_{0} \leqslant\|B\|_{0}$, we have

$$
\|A-B\|_{0} \geqslant\|B\|_{0}-\|A\|_{0} \geqslant \frac{\psi}{1+\psi}\|B\|_{0}
$$

and thus $\frac{\|A\|_{0}+\|B\|_{0}}{\|A-B\|_{0}} \leqslant 2(1+\psi) / \psi$. Hence, $\frac{\|A\|_{0}+\|B\|_{0}}{\|A-B\|_{0}}<4 / \psi$, which yields w.h.p. time $O\left(n+\varepsilon^{-2} \psi^{-1} \log n\right)$.
We execute in parallel the algorithms from Lemma 4.8 and Lemma 4.10. Once the faster algorithm outputs a solution, we terminate the execution of the slower one. Note that this procedure runs w.h.p in time $O\left(\min \left\{\|A\|_{0}, n+\varepsilon^{-2} \psi^{-1} \log n\right\}\right)$, and returns w.h.p. the desired estimator $Y$.

To implement Step 3 of Algorithm 7, we simply apply Theorem 4.11 with $A, \varepsilon$ and $v=z^{(j)}$ to each sampled column $j \in \bigcup_{0 \leqslant i \leqslant \log n+1} C^{(i)}$.

### 4.2.4 Analyzing the Runtime of Algorithm 7

Consider Algorithm 7. In Steps 1, 2 and 2.1, from each of the $O(\log n)$ weight classes we sample $O\left(\varepsilon^{-2} \log n\right)$ columns. In Step 2.2, for each sampled column we use Lemma 4.7, which takes time $O\left(\varepsilon^{-2} n \log m\right)$ per column, or $O\left(\varepsilon^{-4} n \log m \log ^{2} n\right)$ in total. Further, in Step 3, we use Theorem 4.11 for each sampled column, which w.h.p. takes time $O\left(\min \left\{\|A\|_{0}, n+\varepsilon^{-2} \psi^{-1} \log n\right\}\right)$ per column, or in total

$$
O\left(\min \left\{\|A\|_{0}, n+\varepsilon^{-2} \psi^{-1} \log n\right\} \cdot \varepsilon^{-2} \log ^{2} n\right)
$$

Then, the total runtime is bounded by

$$
O\left(\varepsilon^{-4} n \log m \log ^{2} n+\min \left\{\|A\|_{0} \varepsilon^{-2} \log ^{2} n, \varepsilon^{-4} \psi^{-1} \log ^{3} n\right\}\right)
$$

### 4.2.5 Proof of Lemma 1.9

Note that we are given the number of nonzero entries $\ell_{j}=\left\|A_{:, j}\right\|_{0}$ for each column. We want to first sample a column $X \in[n]$ such that $\operatorname{Pr}[X=j]=\ell_{j} / \sum_{k \in[n]} \ell_{k}$, then sample $Y \in\left[\ell_{j}\right]$ uniformly, read $B_{X}[Y]=\left(i, A_{i, X}\right)$, and return $A_{i, X}$. Observe that this process indeed samples each nonzero entry of $A$ with the same probability, since the probability of sampling a particular nonzero entry $(i, j)$ is $\left(\ell_{j} / \sum_{k \in[n]} \ell_{k}\right) \cdot\left(1 / \ell_{j}\right)=1 / \sum_{k \in[n]} \ell_{k}$. Sampling $Y \in\left[\ell_{j}\right]$ uniformly can be done in constant time by assumption. For sampling $X$, we use the classic Alias Method by Walker [Wal74], which is given the probabilities $\operatorname{Pr}[X=1], \ldots, \operatorname{Pr}[X=n]$ as an input and computes, in time $O(n)$, a data structure that allows to sample from $X$ in time $O(1)$. This finishes the construction.

## Part II

## Approximate Spectral Clustering

## Proceedings

This part is based on an article [KM16] that was presented at ESA'16 and on its journal version which is under submission as of July 2018. The full ArXiv version is available online [KM18].

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## Chapter 5

## Introduction

A cluster in an undirected graph $G=(V, E)$ is a subset $S$ of nodes whose volume is large compared to the number of outside connections. Formally, the conductance of $S$ is defined as

$$
\begin{equation*}
\phi(S) \stackrel{\text { def }}{=} \frac{|E(S, \bar{S})|}{\min \{\mu(S), \mu(\bar{S})\}}, \tag{5.1}
\end{equation*}
$$

where the volume of $S$ is given by $\mu(S) \stackrel{\text { def }}{=} \sum_{v \in S} \operatorname{deg}(v)$. We are interested in the problem of partitioning the nodes into a given number $k$ of clusters in a way that (approximately) minimizes the $k$-way conductance

$$
\begin{equation*}
\widehat{\rho}(k) \stackrel{\text { def }}{=} \min _{\text {partition }}^{\left(P_{1}, \ldots, P_{k}\right) \text { of } V} \max _{i \in\{1, \ldots, k\}} \phi\left(P_{i}\right) . \tag{5.2}
\end{equation*}
$$

The 2 -way partitioning constant is also known as the conductance of the graph and is denoted as

$$
\begin{equation*}
\phi_{G} \stackrel{\text { def }}{=} \min _{S \subseteq V} \phi(S) . \tag{5.3}
\end{equation*}
$$

The $k$-way partitioning problem arises in many applications, e.g., image segmentation and exploratory data analysis. We refer to the survey [vL07] for additional information. Further, the surveys [SM00, KVV04, vL07] discuss properties of graphs with small or large conductance.

## Hardness and Approximation

The $k$-way partitioning problem is known to be NP-hard, even for $k=2$ [MS90]. In the case when $k=2$, the $k$-way partitioning problem reduces to the graph conductance problem (5.3), for which there is an approximation algorithm [Chu97] that computes a bipartition $(S, \bar{S})$ such that $\phi(S) \leqslant \sqrt{2 \phi_{G}}$. The algorithm computes an eigenvector corresponding to the second smallest eigenvalue of a normalized Laplacian matrix, sorts the eigenvector's entries, and performs a sweep over the sorted vector. It is guaranteed that one of the resulting sweep sets satisfies the approximation bound.

The fact that the second eigenvector encodes sufficient information for computing an approximate bipartition with small conductance, motivate researchers to consider the bottom $k$ eigenvectors in order to approximately solve the $k$-way partitioning problem. The resulting approach is called Spectral Clustering.

## Spectral Clustering

Given an undirected graph $G=(V, E)$ and a number of clusters $k$, the Spectral Clustering algorithm consists of the following two steps:
(i) Compute the bottom $k$ eigenvectors of the normalized Laplacian matrix of $G$, and store them as the columns of a matrix $Y \in \mathbb{R}^{n \times k}$. The $i$-th node of $G$ is associated with the $i$-th row of $Y$, i.e. a vector in $\mathbb{R}^{k}$. This step is known as Spectral Embedding (SE).
(ii) Partition the resulting vector set into $k$ clusters using a $k$-means clustering algorithm.

This meta algorithm has been successfully applied in practice for solving challenging clustering problems, in the fields of: image segmentation, pattern recognition, data mining, community detection and VLSI design [AY95, SM00, NJW01, MBLS01, BN01, LZ04, ZP04, WS05, vL07, WD12, Tas12, CKC ${ }^{+}$16].

## Approximate Spectral Clustering

Exact computation of Spectral Clustering is expensive due to the following two bottlenecks:
(i) the best algorithm for computing a SE exactly requires time $\Omega\left(n^{\omega}\right)$, cf. [Woo14];
(ii) the $k$-means clustering problem is NP-hard [MNV12].

It is therefore necessary to relax the preceding two problems and to focus on designing approximation schemes for them. Several approximation techniques were developed for Spectral Clustering [Pre81, ST14, YHJ09, CCDL14, FBCM04, PP04, BHH ${ }^{+} 06$, WLRB09, Nys30, WD12, Tas12, LC10, Woo14].

The Power method [LC10, Woo14] is perhaps the most popular technique for computing an approximate SE, due to its simplicity and ease of implementation. Furthermore, this technique was successfully applied for low-rank matrix approximation [Woo14], and it has a worst case convergence guarantee in terms of a principal angle between the space spanned by the approximate and the true eigenvectors [GVL96, Theorem 8.2.4].

Although, the $k$-means clustering problem is NP-hard [MNV12], it admits a polynomial time approximation scheme (PTAS) [KSS04, HK05, FMS07, ORSS13]. However, the best PTAS for computing a $(1+\varepsilon)$ approximation incurs a factor $2^{\text {poly }(k / \varepsilon)}$ in the runtime.

On the other hand, it is folklore that the approximate variant of Spectral Clustering which computes an approximate SE via the Power method, and applies to it an approximate $k$-means clustering algorithm, recovers a good approximation of an optimal $k$-way node partition of $G$ and at the same time runs efficiently (in nearly-linear time).

It is an important task for theory to explain the practical success of Approximate Spectral Clustering, and in particular to resolve the following three questions. In order to state them, we need some notation. Let $Y$ be a SE computed exactly, and $\widetilde{Y}$ be an approximate SE computed via the Power method. Further, let $X(\widetilde{X})$ be an optimal $k$-means clustering partition of the rows of $Y(\widetilde{Y})$. Let $\widetilde{X}_{\alpha}$ be a $k$-way row partition of $\tilde{Y}$, computed by an $\alpha$-approximate $k$-means clustering algorithm. The following questions arise:

Q1. Show that $\widetilde{X_{\alpha}}$ is a good approximation of $X$.
Q2. Show that the $k$-way node partition of $G$ induced by $\widetilde{X_{\alpha}}$, yields a good approximation of an optimal $k$-way node partition of $G$.
Q3. Show that Approximate Spectral Clustering runs efficiently (in nearly-linear time).

## Eigenvalue Gaps and $k$-Way Partitions

Let $0=\lambda_{1} \leqslant \ldots \leqslant \lambda_{n} \leqslant 2$ be the eigenvalues of a normalized Laplacian matrix of $G$. It was observed experimentally [vL07, For10] that a large gap between $\lambda_{k+1}$ and $\lambda_{k}$ guarantees a good $k$-way node partition of $G$ and this was formally proven in [LGT12, GT14]. Lee, Gharan and Trevisan [LGT12] studied the $k$-way expansion constant defined as

$$
\begin{equation*}
\rho(k) \stackrel{\text { def }}{=} \min _{\text {disjoint } S_{1}, \ldots, S_{k}} \max _{i \in\{1, \ldots, k\}} \phi\left(S_{i}\right) \tag{5.4}
\end{equation*}
$$

and related it to $\lambda_{k}$ via higher-order Cheeger inequalities

$$
\begin{equation*}
\lambda_{k} / 2 \leqslant \rho(k) \leqslant O\left(k^{2}\right) \sqrt{\lambda_{k}} . \tag{5.5}
\end{equation*}
$$

For related works on higher-order Cheeger inequalities, we refer the reader to [LRTV12, KLL17]. Gharan and Trevisan [GT14] showed that the $k$-way conductance is at most a factor $k$ away from the $k$-way expansion constant, i.e.,

$$
\begin{equation*}
\rho(k) \leqslant \widehat{\rho}(k) \leqslant k \cdot \rho(k) . \tag{5.6}
\end{equation*}
$$

In particular, (5.5) and (5.6) together yield that $\lambda_{k+1} \gg O\left(k^{3}\right) \sqrt{\lambda_{k}}$ implies $\widehat{\rho}(k+1) \gg \widehat{\rho}(k)$. Thus, there is a $k$-way node partition $\left(P_{1}, \ldots, P_{k}\right)$ of $G$ such that $\phi\left(P_{i}\right) \leqslant O\left(k^{3}\right) \sqrt{\lambda_{k}}$ for all $i$, and simultaneously the best $(k+1)$-way partition is significantly worse.

## Prior Work

Ng et al. [NJW01] reported that ASC performs very well on challenging clustering instances, and initiated the study for finding a formal explanation for the practical success of ASC. Using tools from matrix perturbation theory, they derived sufficient conditions under which the vectors of a SE form tight clusters. However, their analysis does not apply to approximate SEs, and does not give guarantees for the induced $k$-way node partition of $G$.

Peng et al. [PSZ17] showed that for all instances satisfying the eigenvalue gap assumption $\lambda_{k+1} / \widehat{\rho}(k) \geqslant$ $\Omega\left(k^{3}\right)$, any $O(1)$-approximate $k$-means partition of a normalized SE $Y^{\prime}$ induces a good approximation of
an optimal $k$-way node partition of $G$. Notably, their analysis yields the first approximation guarantees in terms of the $k$-way conductance. However, their analysis does not apply to approximate SE, and also computing an $O(1)$-approximation $k$-means partition using any known PTAS [HK05, FMS07, ORSS13] incurs an exponential factor of $2^{\Omega(k)}$ in the running time.

Boutsidis et al. [BKG15] showed that an approximate $k$-means partition of an approximate SE $\widetilde{Y}$ computed via the Power method, yields a $k$-means partition $P$ of the exact SE $Y$ such that the $k$-means cost of $P$ yields an additive approximation to the optimum $k$-means cost of $Y$. This gives an affirmative answer to question Q1. Further, the authors stated as main open problems to resolve questions Q2 and Q3.

Besides designing a PTAS for the $k$-means clustering problems, Ostrovsky et al. [ORSS13] analyzed a variant of Lloyd $k$-means clustering algorithm. They showed that on input a set of $n$ vectors in $\mathbb{R}^{k}$ satisfying a natural well-clusterable assumption, the algorithm efficiently computes a good approximation of an optimal $k$-means partition. In particular, the algorithm runs in time $O\left(k^{2}\left(n+k^{2}\right)\right)$.

A natural question to ask is whether the analysis of Peng et al. [PSZ17], Boutsidis et al. [BKG15] and Ostrovsky et al. [ORSS13] can be integrated and extended to answer the questions Q2 and Q3?

## Our Contribution: An Overview

We give a comprehensive analysis of ASC building on the work of Peng et al. [PSZ17], Boutsidis et al. [BKG15] and Ostrovsky et al. [ORSS13]. We show that the Approximate Spectral Clustering i) runs efficiently, and ii) yields a good approximation of an optimal $k$-way node partition of $G$. Moreover, we strengthen the quality guarantees of a structural result of Peng et al. [PSZ17] by a factor of $k$, and simultaneously weaken the eigenvalue gap assumption. Further, our analysis shows that the Approximate Spectral Clustering finds a $k$-way node partition of $G$ with the strengthened quality guarantees. This gives an affirmative answer to questions Q2 and Q3.

### 5.1 Notation

## $k$-means Clustering Problem

Let $\mathcal{X}$ be a set of vectors of the same dimension. The $k$-means cost of a partition $\left(X_{1}, \ldots, X_{k}\right)$ of $\mathcal{X}$ is given by

$$
\operatorname{Cost}\left(\left\{X_{i}\right\}_{i=1}^{k}\right) \stackrel{\text { def }}{=} \sum_{i=1}^{k} \sum_{x \in X_{i}}\left\|x-c_{i}\right\|_{2}^{2}, \quad \text { where } \quad c_{i} \stackrel{\text { def }}{=} \frac{1}{\left|X_{i}\right|} \sum_{x \in X_{i}} x
$$

is the gravity center of $X_{i}$, for all $i \in\{1, \ldots, k\}$. Then, the optimum $k$-means cost (of clustering $\mathcal{X}$ into $k$ sets) is defined as

$$
\triangle_{k}(\mathcal{X}) \stackrel{\text { def }}{=} \min _{\text {partition }}^{\left(X_{1}, \ldots, X_{k}\right) \text { of } \mathcal{X}} \operatorname{Cost}\left(\left\{X_{i}\right\}_{i=1}^{k}\right)
$$

A $k$-means partition $\left(X_{1}, \ldots, X_{k}\right)$ of $\mathcal{X}$, with corresponding gravity centers $c_{1}, \ldots, c_{k}$ as above, is $\alpha$ approximate if $\operatorname{Cost}\left(\left\{X_{i}\right\}_{i=1}^{k}\right) \leqslant \alpha \cdot \triangle_{k}(\mathcal{X})$. Given a matrix $Y$, we abuse notation and write $\triangle_{k}(Y)$ to denote the optimum $k$-means cost of partitioning the rows of matrix $Y$.

## Spectral Embeddings

Given an undirected graph $G=(V, E)$ with $m=|E|$ edges and $n=|V|$ nodes, let $D$ be the diagonal degree matrix and $A$ be the adjacency matrix. Then, the graph Laplacian matrix is defined as $L=D-A$, and the normalized Laplacian matrix is given by $\mathcal{L}_{G}=I-\mathcal{A}$, where $\mathcal{A}=D^{-1 / 2} A D^{-1 / 2}$. Further, let $f_{i} \in \mathbb{R}^{V}$ be the eigenvector corresponding to the $i$-th smallest eigenvalue $\lambda_{i}$ of $\mathcal{L}_{G}$.

The canonical Spectral Embedding, for short canonical SE, is defined as a matrix $Y \in \mathbb{R}^{n \times k}$ composed of the bottom $k$ eigenvectors ${ }^{1}$ of $\mathcal{L}_{G}$ corresponding to the $k$ smallest eigenvalues. The approximate SE is computed via the Power method ${ }^{2}$. Namely, let $S \in \mathbb{R}^{n \times k}$ be a matrix whose entries are i.i.d. samples from

[^6]the standard Gaussian distribution $N(0,1)$ and $p$ be the number of iterations. Then, the approximate SE $\widetilde{Y}$ is given by:
\[

$$
\begin{equation*}
\text { 1) } M \stackrel{\text { def }}{=} I+\mathcal{A} ; \quad \text { 2) Let } \tilde{U} \widetilde{\Sigma}^{V} \tilde{V}^{\mathrm{T}} \text { be the SVD }{ }^{3} \text { of } M^{p} S ; \quad \text { and } \quad \text { 3) } \widetilde{Y} \stackrel{\text { def }}{=} \widetilde{U} \in \mathbb{R}^{n \times k} \text {. } \tag{5.7}
\end{equation*}
$$

\]

Peng et al. [PSZ17] do not apply $k$-means directly to the canonical SE, but first normalize it by dividing the row corresponding to $u$ by $\sqrt{d(u)}$ and then put $d(u)$ copies of the resulting vector into the $k$-means clustering instance. This repetition of vectors is crucial for their analysis, in order to achieve approximation guarantees in terms of volume overlap and conductance. We follow their approach.

We construct a matrix $Y^{\prime} \in \mathbb{R}^{2 m \times k}$ such that for every node $u \in V$, we insert $d(u)$ many copies of the normalized row $Y(u,:) / \sqrt{d(u)}$ to $Y^{\prime}$. Formally, the normalized $\mathrm{SE} Y^{\prime}$ and the approximate normalized SE $\widetilde{Y^{\prime}}$ are defined by

$$
Y^{\prime} \stackrel{\text { def }}{=}\left(\begin{array}{c}
\mathbf{1}_{d(1)} \frac{Y(1,:)}{\sqrt{d(1)}}  \tag{5.8}\\
\cdots \\
\mathbf{1}_{d(n)} \frac{Y(n,:)}{\sqrt{d(n)}}
\end{array}\right)_{2 m \times k} \quad \text { and } \quad \widetilde{Y^{\prime}} \stackrel{\text { def }}{=}\left(\begin{array}{c}
\mathbf{1}_{d(1)} \frac{\widetilde{Y}(1,:)}{\sqrt{d(1)}} \\
\cdots \\
\mathbf{1}_{d(n)} \frac{\widetilde{Y}(n,:)}{\sqrt{d(n)}}
\end{array}\right)_{2 m \times k},
$$

where $\mathbf{1}_{d(i)}$ is the all-one column vector with dimension $d(i)$.
We can assume w.l.o.g. that a $k$-means clustering algorithm applied on $Y^{\prime}\left(\widetilde{Y^{\prime}}\right)$, outputs a $k$-means partition such that all copies of row $Y(v,:) / \sqrt{d(v)}(\widetilde{Y}(v,:) / \sqrt{d(v)})$ belong to the same cluster, for all nodes $v$. Thus, the algorithm induces a $k$-way node partition of $G$.

### 5.1.1 Prior Structural Result

A key prior structural result, established by Peng et al. [PSZ17], connects the normalized SE $Y^{\prime}$, $\alpha-$ approximate $k$-means clustering, the $k$-way conductance $\hat{\rho}(k)$, as given in Equation (5.2), and the $(k+1)$-st eigenvalue $\lambda_{k+1}$ of the normalized Laplacian matrix $\mathcal{L}_{G}$. In particular, they proved the following statement under a gap assumption defined in terms of

$$
\Upsilon \stackrel{\text { def }}{=} \frac{\lambda_{k+1}}{\widehat{\rho}(k)}
$$

Theorem 5.1. [PSZ17, Theorem 1.2] Let $k \geqslant 3$ and $G$ be a graph satisfying the gap assumption ${ }^{4}$

$$
\begin{equation*}
\delta \stackrel{\text { def }}{=} 2 \cdot 10^{5} \cdot k^{3} / \Upsilon \leqslant 1 / 2 \tag{5.9}
\end{equation*}
$$

Let $\left(P_{1}, \ldots, P_{k}\right)$ be a $k$-way node partition of $G$ achieving $\widehat{\rho}(k)$, and let $\left(A_{1}, \ldots, A_{k}\right)$ be the $k$-way node partition of $G$ induced by an $\alpha$-approximate $k$-means partition of the normalized $S E Y^{\prime}$. Then, for every $i \in\{1, \ldots, k\}$ it hold (after suitable renumbering of one of the partitions) that

$$
\text { 1) } \mu\left(A_{i} \triangle P_{i}\right) \leqslant \alpha \delta \cdot \mu\left(P_{i}\right) \quad \text { and } \quad \text { 2) } \phi\left(A_{i}\right) \leqslant(1+2 \alpha \delta) \cdot \phi\left(P_{i}\right)+2 \alpha \delta \text {. }
$$

Under a stronger eigenvalue gap assumption $2 \cdot 10^{5} \cdot k^{5} / \Upsilon \leqslant 1 / 2$, Peng et al. [PSZ17] gave an algorithm that finds in time $O(m \cdot \operatorname{poly} \log (n))$ a $k$-way node partition of $G$ with essentially the guarantees stated in Theorem 5.1. However, their algorithmic result substitutes normalized SE with Heat Kernel Embedding and $k$-means clustering with locality sensitive hashing. Thus, the algorithmic part of their paper does not explain the success of Approximate Spectral Clustering.

### 5.1.2 Our Contribution

We give affirmative answer to the questions Q2 and Q3. On the way, we also strengthen the approximation guarantees in Theorem 5.1 by a factor of $k$ and simultaneously weaken the eigenvalue gap assumption.

Let $\mathcal{P}$ be the set of all $k$-way partitions $\left(P_{1}, \ldots, P_{k}\right)$ achieving the $k$-way conductance $\widehat{\rho}(k)$. Let

$$
\widehat{\rho}_{\mathrm{avr}}(k) \stackrel{\text { def }}{=} \min _{\left(P_{1}, \ldots, P_{k}\right) \in \mathcal{P}} \frac{1}{k} \sum_{i=1}^{k} \phi\left(P_{i}\right)
$$

[^7]be the minimum average conductance over all $k$-way partitions in $\mathcal{P}$. Note that $\widehat{\rho}_{\text {avr }}(k) \leqslant \widehat{\rho}(k)$. Our gap assumption is defined in terms of
$$
\Psi \stackrel{\text { def }}{=} \frac{\lambda_{k+1}}{\widehat{\rho}_{\mathrm{avr}}(k)} .
$$

In the remainder, we denote by $\left(P_{1}, \ldots, P_{k}\right)$ a $k$-way node partition of $G$ achieving $\widehat{\rho}_{\text {avr }}(k)$.
We now present our main result, consisting of a structural and an algorithmic statement.
Theorem 5.2. a) (Existence of a Good Clustering) Let $k \geqslant 3$ and $G$ be a graph satisfying

$$
\begin{equation*}
\delta \stackrel{\text { def }}{=} 20^{4} \cdot k^{3} / \Psi \leqslant 1 / 2 \tag{5.10}
\end{equation*}
$$

Let $\left(P_{1}, \ldots, P_{k}\right)$ be a $k$-way node partition of $G$ achieving $\widehat{\rho}_{\text {avr }}(k)$, and let $\left(A_{1}, \ldots, A_{k}\right)$ be the $k$-way node partition of $G$ induced by an $\alpha$-approximate $k$-means partition of the normalized $S E Y^{\prime}$. Then, for every $i \in\{1, \ldots, k\}$ it hold (after suitable renumbering of one of the partitions) that

$$
\text { 1) } \left.\mu\left(A_{i} \triangle P_{i}\right) \leqslant \frac{\alpha \delta}{10^{3} k} \cdot \mu\left(P_{i}\right) \quad \text { and } \quad 2\right) \phi\left(A_{i}\right) \leqslant\left(1+\frac{2 \alpha \delta}{10^{3} k}\right) \cdot \phi\left(P_{i}\right)+\frac{2 \alpha \delta}{10^{3} k} \text {. }
$$

b) (An Efficient Algorithm) If in addition $k / \delta \geqslant 10^{9}$, then the variant of Lloyd algorithm analyzed by Ostrovsky et al. [ORSS13] when applied to the approximate normalized $S E \widetilde{Y^{\prime}}$, induces in time $O\left(m\left(k^{2}+\right.\right.$ $\left.\left.\frac{\ln n}{\lambda_{k+1}}\right)\right)$ with constant probability a $k$-way node partition $\left(A_{1}, \ldots, A_{k}\right)$ of $G$ such that for every $i \in\{1, \ldots, k\}$ it hold (after suitable renumbering of one of the partitions) that

$$
\text { 3) } \mu\left(A_{i} \triangle P_{i}\right) \leqslant \frac{2 \delta}{10^{3} k} \cdot \mu\left(P_{i}\right) \quad \text { and } \quad \text { 4) } \phi\left(A_{i}\right) \leqslant\left(1+\frac{4 \delta}{10^{3} k}\right) \cdot \phi\left(P_{i}\right)+\frac{4 \delta}{10^{3} k}
$$

Part (a) of Theorem 5.2 strengthens the quality guarantees in Theorem 5.1 by a factor of $k$, and simultaneously weaken the eigenvalue gap assumption. Part (b) of Theorem 5.2 gives a comprehensive analysis of Approximate Spectral Clustering, and demonstrates that the algorithm i) runs efficiently, and ii) yields a good approximation of an optimal $k$-way node partition of $G$.

Further, it shows that the Approximate Spectral Clustering finds a $k$-way node partition of $G$ with the strengthened quality guarantees, and whenever $k \leqslant(\log n)^{O(1)}$ and $\lambda_{k+1} \geqslant 1 /(\log n)^{O(1)}$, the algorithm runs in nearly linear time. This answers affirmatively questions Q2 and Q3.

## Remarks

The variant of Lloyd $k$-means clustering algorithm, analyzed by Ostrovsky et al. [ORSS13], is efficient only for inputs $\mathcal{X}$ satisfying $\triangle_{k}(\mathcal{X}) \leqslant \varepsilon^{2} \triangle_{k-1}(\mathcal{X})$ for some $\varepsilon \in\left(0, \varepsilon_{0}\right]$, where $\varepsilon_{0}=6 / 10^{7}$. The authors stated that their result should also hold for a larger $\varepsilon_{0}$, and mentioned that they did not attempt to maximize $\varepsilon_{0}$.

An anonymous reviewer of the conference version of this paper, suggested to include a numerical example. Consider a graph which consists of $k$ cliques each of size $n / k^{5}$, plus $k$ additional edges that connect the cliques in the form of a ring such that no two edges of the ring share a vertex. This graph is a trivial clustering instance, and for any constant $k$ it holds that $\lambda_{k+1} \geqslant 1-k / n$, see Theorem 7.17. Observe that $\hat{\rho}_{\text {avr }}(k)=\widehat{\rho}(k) \approx(k / n)^{2}$. For the gap assumption to hold we need $\lambda_{k+1} \geqslant 2 \cdot 20^{4} \cdot k^{3} \cdot \widehat{\rho}_{\text {avr }}(k)$. This implies $n \geqslant \sqrt{2 \cdot 20^{4} \cdot k^{5} / \lambda_{k+1}}$. For small $k$, this is a modest requirement on the size of the graph.

For the algorithmic result, we need in addition $\delta \leqslant k \cdot \varepsilon_{0} / 600$. For the gap condition to hold, we need $\lambda_{k+1} \geqslant\left(600 / \varepsilon_{0} k\right) \cdot 20^{4} \cdot k^{3} \cdot\left(k^{2} / n^{2}\right)$ or $n \geqslant \sqrt{20^{6} \cdot k^{4} /\left(\varepsilon_{0} \lambda_{k+1}\right)}$. For $\varepsilon_{0}=6 / 10^{7}$, this amounts to $n \geqslant \sqrt{2^{4} \cdot 10^{13} \cdot k^{4} / \lambda_{k+1}}$, a quite large lower bound on $n$.

The statement that Part (b) of Theorem 5.2 gives a theoretical support for the practical success of Approximate Spectral Clustering, therefore has to be taken with a grain of salt. It is only an asymptotic statement and does not explain the good behavior on small graphs.

[^8]
### 5.2 Our Techniques

In Chapter 6, we give a refined spectral analysis of [PSZ17] which yields the improved structural result in Part (a) of Theorem 5.2. In Chapter 7, we connect Part (a) of Theorem 5.2 with the work of Ostrovsky et al. [ORSS13] and Boutsidis et al. [BKG15], yielding the algorithmic result in Part (b) of Theorem 5.2.

Ostrovsky et al. [ORSS13] analyzed a variant of Lloyd $k$-means clustering algorithm. We refer to this algorithm as the ORSS clustering algorithm. The ORSS-algorithm is efficient only for inputs $\mathcal{X}$ satisfying: some partition into $k$ clusters is much better than any partition into $k-1$ clusters. Formally, it states

Theorem 5.3. [ORSS13, Theorem 4.15] Assuming that $\triangle_{k}(\mathcal{X}) \leqslant \varepsilon^{2} \cdot \triangle_{k-1}(\mathcal{X})$ for $\varepsilon \in\left(0,6 \cdot 10^{-7}\right]$, the ORSS-algorithm runs in time $O\left(n k d+k^{3} d\right)$ and returns with probability at least $1-O(\sqrt{\varepsilon})$ a $k$-way partition of $\mathcal{X}$ with cost at most $\left[\left(1-\varepsilon^{2}\right) /\left(1-37 \varepsilon^{2}\right)\right] \triangle_{k}(\mathcal{X})$.

Let $Z \in \mathbb{R}^{n \times k}$ be a matrix and $\left(R_{1}, \ldots, R_{k}\right)$ be a row partition of $Z$. Let $c_{j}=\frac{1}{\left|R_{j}\right|} \sum_{u \in R_{j}} Z_{u, \text { : }}$ be the gravity center of cluster $R_{j}$, for all $j \in\{1, \ldots, k\}$. We next express in matrix notation the $k$-means cost of partition $\left(R_{1}, \ldots, R_{k}\right)$. To this end, we introduce an indicator matrix $X \in \mathbb{R}^{n \times k}$ such that $X_{i j}=1 / \sqrt{\left|R_{j}\right|}$ if row $Z_{i, \text { : }}$ belongs to cluster $R_{j}$, and $X_{i j}=0$ otherwise. Then, $\left(X X^{T} Z\right)_{i,:}=c_{j}$, where row $Z_{i,:}$ belongs to cluster $R_{j}$. Hence, the $k$-means cost of $\left(R_{1}, \ldots, R_{k}\right)$ becomes

$$
\begin{equation*}
\operatorname{Cost}\left(\left\{R_{i}\right\}_{i=1}^{k}\right)=\sum_{j=1}^{k} \sum_{u \in R_{j}}\left\|Z_{u,:} c_{j}\right\|_{2}^{2}=\left\|Z-X X^{\mathrm{T}} Z\right\|_{F}^{2} \tag{5.11}
\end{equation*}
$$

## Our Analytical Approach

Our main technical contribution is to prove that the approximate normalized $\mathrm{SE} \widetilde{Y^{\prime}}$ computed via the Power method is $\varepsilon$-separated, i.e. the assumption $\triangle_{k}\left(\widetilde{Y^{\prime}}\right)<\varepsilon^{2} \cdot \triangle_{k-1}\left(\widetilde{Y^{\prime}}\right)$ of Ostrovsky et al. [ORSS13] is satisfied. This implies, by Theorem 5.3, that the ORSS-algorithm runs efficiently on $\widetilde{Y^{\prime}}$. Let the resulting $k$-way row partition of $\widetilde{Y^{\prime}}$ be encoded by the indicator matrix $\widetilde{X^{\prime}}$.

Then, building on the work of [BM14, BKG15], we show that $\widetilde{X^{\prime}}$ is a good approximation of an optimal $k$-means partition of the corresponding normalized SE $Y^{\prime}$. Further, using our strengthened structural result in Part (a) of Theorem 5.2, we show that $\widetilde{X^{\prime}}$ induces a good approximation of an optimal $k$-way node partition of graph $G$, in terms of volume overlap and conductance.

First, we establish in Section 7.1 the assumption of Ostrovsky et al. [ORSS13] for the normalized SE $Y^{\prime}$.
Theorem 5.4. (normalized $S E$ is $\varepsilon$-separated) Let $G$ be a graph that satisfies $\Psi=20^{4} \cdot k^{3} / \delta, \delta \in(0,1 / 2]$ and $k / \delta \geqslant 10^{9}$. Then for $\varepsilon=6 \cdot 10^{-7}$ it holds $\triangle_{k}\left(Y^{\prime}\right) \leqslant \varepsilon^{2} \cdot \triangle_{k-1}\left(Y^{\prime}\right)$.

Theorem 5.4 does not suffice for proving Part (b) of Theorem 5.2, since it requires the analogous statement for the approximate normalized $\mathrm{SE} \widetilde{Y^{\prime}}$.

In Subsection 7.2.2, we show that an $\alpha$-approximate $k$-means clustering algorithm applied to the approximate normalized $\mathrm{SE}{\widetilde{Y^{\prime}}}^{\prime}$, yields an approximate $k$-way row partition of the corresponding normalized SE $Y^{\prime}$.

Theorem 5.5. (Similar to [BKG15, Theorem 6], but analyzes the approximate normalized SE) Let $\varepsilon, \delta_{p} \in(0,1)$ be arbitrary. Compute the approximate normalized $S E \widetilde{Y^{\prime}}$ via the Power method with $p \geqslant \ln \left(8 n k / \varepsilon \delta_{p}\right) / \ln \left(1 / \gamma_{k}\right)$ iterations and $\gamma_{k}=\left(2-\lambda_{k+1}\right) /\left(2-\lambda_{k}\right)<1$. Run on the rows of $\widetilde{Y^{\prime}}$ an $\alpha$-approximate $k$-means clustering algorithm with failure probability $\delta_{\alpha}$. Let the outcome be a clustering indicator matrix $\widetilde{X_{\alpha}^{\prime}} \in \mathbb{R}^{n \times k}$. Then, with probability at least $1-2 e^{-2 n}-3 \delta_{p}-\delta_{\alpha}$, it holds that

$$
\left\|Y^{\prime}-\widetilde{X_{\alpha}^{\prime}}\left(\widetilde{X_{\alpha}^{\prime}}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F}^{2} \leqslant(1+4 \varepsilon) \cdot \alpha \cdot \triangle_{k}\left(Y^{\prime}\right)+4 \varepsilon^{2}
$$

In Subsection 7.2.3, using Theorem 5.4 and Theorem 5.5, we show that the approximate normalized SE $\widetilde{Y}^{\prime}$ satisfies the assumption of Ostrovsky et al. [ORSS13].
Theorem 5.6. (approximate normalized SE is $\varepsilon$-separated) Assume $\Psi=20^{4} \cdot k^{3} / \delta$ and $k / \delta \geqslant 10^{9}$ for some $\delta \in(0,1 / 2]$. Compute the approximate normalized SE $\widetilde{Y^{\prime}}$ via the Power method with $p \geqslant \Omega\left(\frac{\ln n}{\lambda_{k+1}}\right)$ iterations. Then, for $\varepsilon=6 \cdot 10^{-7}$ it holds with high probability that $\triangle_{k}\left(\widetilde{Y^{\prime}}\right)<5 \varepsilon^{2} \cdot \triangle_{k-1}\left(\widetilde{Y^{\prime}}\right)$.

Finally, in Subsection 7.3, we prove Part (b) of Theorem 5.2 by combining Part (a) of Theorem 5.2, Theorem 5.3, Theorem 5.5 and Theorem 5.6.

## Chapter 6

## Improved Structural Result

### 6.1 Notation

We use the notation adopted in [PSZ17]. Let $\lambda_{j}$ be the $j$-th eigenvalue of the normalized Laplacian matrix $\mathcal{L}_{G}$, and let $f_{j} \in \mathbb{R}^{V}$ be the associated eigenvector $\left(\mathcal{L}_{G} f_{j}=\lambda_{j} f_{j}\right)$.

Let $\overline{g_{i}}=\frac{D^{1 / 2} \chi_{P_{i}}}{\left\|D^{1 / 2} \chi_{P_{i}}\right\|_{2}}$, where $\chi_{P_{i}}$ is the characteristic vector of the subset $P_{i} \subseteq V$. Note that $\overline{g_{i}}$ is the normalized characteristic vector of $P_{i}$ and $\left\|D^{1 / 2} \chi_{P_{i}}\right\|_{2}^{2}=\sum_{v \in P_{i}} d(v)=\mu\left(P_{i}\right)$. The Rayleigh quotient is defined by and satisfies

$$
\mathcal{R}\left(\overline{g_{i}}\right) \stackrel{\text { def }}{=} \frac{\bar{g}_{i}^{\mathrm{T}} \mathcal{L}_{G} \overline{g_{i}}}{\overline{g_{i}} \mathrm{~T} \overline{g_{i}}}=\frac{1}{\mu\left(P_{i}\right)} \chi_{P_{i}}^{\mathrm{T}} L \chi_{P_{i}}=\frac{|E(S, \bar{S})|}{\mu\left(P_{i}\right)}=\phi\left(P_{i}\right),
$$

where the Laplacian matrix $L=D-A$ and the normalized Laplacian matrix $\mathcal{L}_{G}=D^{-1 / 2} L D^{-1 / 2}$.
The eigenvectors $\left\{f_{i}\right\}_{i=1}^{n}$ form an orthonormal basis of $\mathbb{R}^{n}$. Thus each characteristic vector $\overline{g_{i}}$ can be expressed as $\overline{g_{i}}=\sum_{j=1}^{n} \alpha_{j}^{(i)} f_{j}$ for all $i \in\{1, \ldots, k\}$. We define its projection onto the first $k$ eigenvectors by $\widehat{f}_{i}=\sum_{j=1}^{k} \alpha_{j}^{(i)} f_{j}$.

Peng et al. [PSZ17] proved that if the gap parameter $\Upsilon$ is large enough then $\operatorname{span}\left(\left\{\widehat{f}_{i}\right\}_{i=1}^{k}\right)=$ $\operatorname{span}\left(\left\{f_{i}\right\}_{i=1}^{k}\right)$ and the first $k$ eigenvectors can be expressed by $f_{i}=\sum_{j=1}^{k} \beta_{j}^{(i)} \widehat{f}_{j}$, for all $i \in\{1, \ldots, k\}$. Moreover, they demonstrated that each vector $\widehat{g_{i}}=\sum_{j=1}^{k} \beta_{j}^{(i)} \overline{g_{j}}$ approximates the eigenvector $f_{i}$, for all $i \in\{1, \ldots, k\}$. We will show that similar statements hold with weakened gap parameter $\Psi$.

The estimation centers induced by the canonical SE are given by

$$
\begin{equation*}
p^{(i)}=\frac{1}{\sqrt{\mu\left(P_{i}\right)}}\left(\beta_{i}^{(1)}, \ldots, \beta_{i}^{(k)}\right)^{\mathrm{T}} . \tag{6.1}
\end{equation*}
$$

Our analysis crucially relies on spectral properties of the following two matrices. Let $F, B \in \mathbb{R}^{k \times k}$ be square matrices defined by

$$
\begin{equation*}
F_{j, i}=\alpha_{j}^{(i)} \quad \text { and } \quad B_{j, i}=\beta_{j}^{(i)} \tag{6.2}
\end{equation*}
$$

In Figure 6.1, we show the relation among the vectors $f_{i}, \widehat{f_{i}}, \widehat{g_{i}}$ and $\overline{g_{i}}$.


Figure 6.1: The vectors $\left\{f_{i}\right\}_{i=1}^{n}$ are eigenvectors of the normalized Laplacian matrix $\mathcal{L}_{G}$. The vectors $\left\{\bar{g}_{i}\right\}_{i=1}^{k}$ are the normalized characteristic vectors of an optimal partition $\left(P_{1}, \ldots, P_{k}\right)$. For each $i \in$ $\{1, \ldots, k\}$ the vector $\widehat{f}_{i}$ is the projection of vector $\overline{g_{i}}$ onto $\operatorname{span}\left(f_{1}, \ldots, f_{k}\right)$. The vectors $\widehat{f}_{i}$ and $\bar{g}_{i}$ are close for $i \in\{1, \ldots, k\}$. If $\Psi>4 \cdot k^{3 / 2}$, then $\operatorname{span}\left(f_{1}, \ldots, f_{k}\right)=\operatorname{span}\left(\widehat{f_{1}}, \ldots, \widehat{f_{k}}\right)$ and thus we can write $f_{i}=\sum_{j=1}^{k} \beta_{j}^{(i)} \widehat{f_{j}}$. Further, the vectors $f_{i}$ and $\widehat{g_{i}}=\sum_{j=1}^{k} \beta_{j}^{(i)} \overline{g_{j}}$ are close for $i \in\{1, \ldots, k\}$.

### 6.2 Technical Insights

The analysis of Part (a) of Theorem 5.2 follows the proof approach in [PSZ17, Theorem 1.2], but improves upon it in essential ways.

Our first technical insight is that the matrices $B^{\mathrm{T}} B$ and $B B^{\mathrm{T}}$ are close to the identity matrix. We prove this in two steps. In Section 6.4, we show that the vectors $\widehat{g}_{i}$ and $f_{i}$ are close, and then in Section 6.5 we analyze the column space and row space of matrix $B$.

Theorem 6.1 (Matrix $B B^{\mathrm{T}}$ is Close to Identity Matrix). If $\Psi \geqslant 10^{4} \cdot k^{3} / \varepsilon^{2}$ and $\varepsilon \in(0,1)$ then for all distinct $i, j \in\{1, \ldots, k\}$ it holds

$$
1-\varepsilon \leqslant\left\langle B_{i,:}, B_{i,:}\right\rangle \leqslant 1+\varepsilon \quad \text { and } \quad\left|\left\langle B_{i,:}, B_{j,:}\right\rangle\right| \leqslant \sqrt{\varepsilon} .
$$

Using Theorem 6.1, we give a strengthened version of [PSZ17, Lemma 4.2] that depends on the weaken gap parameter $\Psi$.

Lemma 6.2. If $\Psi=20^{4} \cdot k^{3} / \delta$ for some $\delta \in(0,1]$ then for every $i \in\{1, \ldots, k\}$ it holds that

$$
(1-\sqrt{\delta} / 4) \frac{1}{\mu\left(P_{i}\right)} \leqslant\left\|p^{(i)}\right\|_{2}^{2} \leqslant(1+\sqrt{\delta} / 4) \frac{1}{\mu\left(P_{i}\right)} .
$$

Proof. By definition $p^{(i)}=\frac{1}{\sqrt{\mu\left(P_{i}\right)}} \cdot B_{i,:}$ and Theorem 6.1 yields $\left\|B_{i,:}\right\|_{2}^{2} \in[1 \pm \sqrt{\delta} / 4]$.
Using Theorem 6.1 and Lemma 6.2, we establish a strengthened version of [PSZ17, Lemma 4.3] that depends on the weaken gap parameter $\Psi$, and simultaneously shows that the $\ell_{2}$ distance between estimation centers is larger by a factor of $k$.

Lemma 6.3 (Larger Distance Between Estimation Centers). If $\Psi=20^{4} \cdot k^{3} / \delta$ for some $\delta \in\left(0, \frac{1}{2}\right]$ then for any distinct $i, j \in\{1, \ldots, k\}$ it holds that

$$
\left\|p^{(i)}-p^{(j)}\right\|_{2}^{2} \geqslant\left[2 \cdot \min \left\{\mu\left(P_{i}\right), \mu\left(P_{j}\right)\right\}\right]^{-1}
$$

Proof. Since $p^{(i)}$ is a row of matrix $B$, Theorem 6.1 with $\varepsilon=\sqrt{\delta} / 4$ yields

$$
\left\langle\frac{p^{(i)}}{\left\|p^{(i)}\right\|_{2}}, \frac{p^{(j)}}{\left\|p^{(j)}\right\|_{2}}\right\rangle=\frac{\left\langle B_{i,:}, B_{j,:}\right\rangle}{\left\|B_{i,:}\right\|_{2}\left\|B_{j,:}\right\|_{2}} \leqslant \frac{\sqrt{\varepsilon}}{1-\varepsilon}=\frac{2 \delta^{1 / 4}}{3} .
$$

W.l.o.g. assume that $\left\|p^{(i)}\right\|_{2}^{2} \geqslant\left\|p^{(j)}\right\|_{2}^{2}$, say $\left\|p^{(j)}\right\|_{2}=\alpha \cdot\left\|p^{(i)}\right\|_{2}$ for some $\alpha \in(0,1]$. Then by Lemma 6.2 we have $\left\|p^{(i)}\right\|_{2}^{2} \geqslant(1-\sqrt{\delta} / 4) \cdot\left[\min \left\{\mu\left(P_{i}\right), \mu\left(P_{j}\right)\right\}\right]^{-1}$, and hence

$$
\begin{aligned}
\left\|p^{(i)}-p^{(j)}\right\|_{2}^{2} & =\left\|p^{(i)}\right\|_{2}^{2}+\left\|p^{(j)}\right\|_{2}^{2}-2\left\langle\frac{p^{(i)}}{\left\|p^{(i)}\right\|_{2}}, \frac{p^{(j)}}{\left\|p^{(j)}\right\|_{2}}\right\rangle\left\|p^{(i)}\right\|_{2}\left\|p^{(j)}\right\|_{2} \\
& \geqslant\left(\alpha^{2}-\frac{4 \delta^{1 / 4}}{3} \cdot \alpha+1\right)\left\|p^{(i)}\right\|_{2}^{2} \geqslant\left[2 \cdot \min \left\{\mu\left(P_{i}\right), \mu\left(P_{j}\right)\right\}\right]^{-1}
\end{aligned}
$$

Using Lemma 6.2 and Lemma 6.3, the observation that $\Upsilon$ can be replaced by $\Psi$ in all statements in [PSZ17] is technically easy.

Our second technical contribution is to show that the larger $\ell_{2}$ distance between estimation centers, in Lemma 6.3, strengthens [PSZ17, Lemma 4.5] by a factor of $k$. Before we state our result, we need some notation.

The normalized Spectral Embedding map $\mathcal{F}: V \rightarrow \mathbb{R}^{k}$ is defined by

$$
\mathcal{F}(v) \stackrel{\text { def }}{=} \frac{1}{\sqrt{d(v)}}\left(f_{1}(v), \ldots, f_{k}(v)\right)^{\mathrm{T}}=\frac{1}{\sqrt{d(v)}} \cdot[Y(v,:)]^{\mathrm{T}}
$$

for every node $v \in V$. Recall that the normalized $\mathrm{SE} Y^{\prime}$ contains duplicate rows, namely, $d(u)$ many copies of $\mathcal{F}(u)$ for each node $u \in V$.

Suppose an $\alpha$-approximate $k$-means clustering algorithm outputs a $k$-way row partition $\left(R_{1}, \ldots, R_{k}\right)$ of $Y^{\prime}$. We can assume w.l.o.g. that all identical rows of $Y^{\prime}$ are assigned to same cluster, and thus $\left(R_{1}, \ldots, R_{k}\right)$ induces a $k$-way node partition $\left(A_{1}, \ldots, A_{k}\right)$ of $G$. For an arbitrary point set $c_{1}, \ldots, c_{k}$ in $\mathbb{R}^{k}$, we abuse the notation and denote the $k$-means cost of a tuple $\left\{A_{i}, c_{i}\right\}_{i=1}^{k}$ by

$$
\begin{equation*}
\operatorname{Cost}\left(\left\{A_{i}, c_{i}\right\}_{i=1}^{k}\right)=\sum_{i=1}^{k} \sum_{u \in A_{i}} d(u)\left\|\mathcal{F}(u)-c_{i}\right\|_{2}^{2} \tag{6.3}
\end{equation*}
$$

When each point $c_{j}=\frac{1}{\mu\left(A_{j}\right)} \sum_{u \in A_{j}} d(u) \mathcal{F}(u)$ is the gravity center of cluster $R_{j}$, for brevity we write $\operatorname{Cost}\left(\left\{A_{i}\right\}_{i=1}^{k}\right)$ to denote the $k$-means cost of tuple $\left\{A_{i}, c_{i}\right\}_{i=1}^{k}$.
Lemma 6.4 (Volume Overlap). Let $\left(P_{1}, \ldots, P_{k}\right)$ and $\left(A_{1}, \ldots, A_{k}\right)$ be $k$-way node partitions of $G$. Suppose for every permutation $\pi:\{1, \ldots, k\} \rightarrow\{1, \ldots, k\}$ there is an index $i \in\{1, \ldots, k\}$ such that

$$
\begin{equation*}
\mu\left(A_{i} \triangle P_{\pi(i)}\right) \geqslant \frac{2 \varepsilon}{k} \cdot \mu\left(P_{\pi(i)}\right) \tag{6.4}
\end{equation*}
$$

where $\varepsilon \in(0,1)$ is a parameter. If $\Psi=20^{4} \cdot k^{3} / \delta$ for some $\delta \in\left(0, \frac{1}{2}\right]$, and $\varepsilon \geqslant 64 \alpha \cdot k^{3} / \Psi$ then

$$
\operatorname{Cost}\left(\left\{A_{i}\right\}_{i=1}^{k}\right)>\frac{2 \alpha k^{2}}{\Psi}
$$

With the above lemmas in place, the proof of Part (a) of Theorem 5.2 is then completed as in [PSZ17]. For completeness, we present the proof.

### 6.3 Proof of Improved Structural Result

In this Section, we prove Part (a.1) of Theorem 5.2. Crucial to our analysis is the following result, which we prove in the next Section 6.4, showing that vectors $\widehat{g}_{i}$ and $f_{i}$ are close, c.f. Figure 6.1.
Theorem 6.5. If $\Psi>4 \cdot k^{3 / 2}$, then for every $i \in\{1, \ldots, k\}$ the vectors $f_{i}$ and $\widehat{g_{i}}=\sum_{j=1}^{k} \beta_{j}^{(i)} \overline{g_{j}}$ satisfy

$$
\left\|f_{i}-\widehat{g_{i}}\right\|^{2} \leqslant\left(1+\frac{3 k}{\Psi}\right) \cdot \frac{k}{\Psi}
$$

Lemma $6.6\left(\left(P_{1}, \ldots, P_{k}\right)\right.$ is a good $k$-means partition). If $\Psi>4 \cdot k^{3 / 2}$, then there are vectors $\left\{p^{(i)}\right\}_{i=1}^{k}$ such that

$$
\operatorname{Cost}\left(\left\{P_{i}, p^{(i)}\right\}_{i=1}^{k}\right) \leqslant\left(1+\frac{3 k}{\Psi}\right) \cdot \frac{k^{2}}{\Psi}
$$

Proof. Let $\left(P_{1}, \ldots, P_{k}\right)$ be a $k$-way node partition of $G$ achieving $\widehat{\rho}_{\text {avr }}(k)$. Peng et al. [PSZ17, Lemma 4.1] showed that $\operatorname{Cost}\left(\left\{P_{i}, p^{(i)}\right\}_{i=1}^{k}\right)=\sum_{j=1}^{k}\left\|f_{j}-\widehat{g}_{j}\right\|_{2}^{2}$, and thus the statement follows by Theorem 6.5.

For completeness, we now prove the preceding equation. By definition, $p_{j}^{(i)}=\beta_{i}^{(j)} / \sqrt{\mu\left(P_{i}\right)}$ and $\widehat{g}_{j}=\sum_{i=1}^{k} \beta_{i}^{(j)} \cdot \frac{D^{1 / 2} \chi_{P_{i}}}{\sqrt{\mu\left(P_{i}\right)}}$, where $\chi_{P_{i}}$ is characteristic vector of the node subset $P_{i}$. Then,

$$
\begin{gathered}
\operatorname{Cost}\left(\left\{P_{i}, p^{(i)}\right\}_{i=1}^{k}\right)=\sum_{i=1}^{k} \sum_{u \in P_{i}} d(u)\left\|\mathcal{F}(u)-p^{(i)}\right\|_{2}^{2} \\
=\sum_{j=1}^{k} \sum_{i=1}^{k} \sum_{u \in P_{i}}\left(f_{j}(u)-\frac{\sqrt{d(u)}}{\sqrt{\mu\left(P_{i}\right)}} \beta_{i}^{(j)}\right)^{2}=\sum_{j=1}^{k}\left\|f_{j}-\widehat{g}_{j}\right\|_{2}^{2} .
\end{gathered}
$$

Lemma 6.7 (Only partitions close to $\left(P_{1}, \ldots, P_{k}\right)$ are good). Under the hypothesis of Theorem 5.2, the following holds. If for every permutation $\sigma:\{1, \ldots, k\} \rightarrow\{1, \ldots, k\}$ there exists an index $i \in\{1, \ldots, k\}$ such that

$$
\begin{equation*}
\mu\left(A_{i} \triangle P_{\sigma(i)}\right) \geqslant \frac{8 \alpha \delta}{10^{4} k} \cdot \mu\left(P_{\sigma(i)}\right) \tag{6.5}
\end{equation*}
$$

Then it holds that

$$
\begin{equation*}
\operatorname{Cost}\left(\left\{A_{i}\right\}_{i=1}^{k}\right)>\frac{2 \alpha k^{2}}{\Psi} \tag{6.6}
\end{equation*}
$$

We note that Lemma 6.7 follows directly by applying Lemma 6.4 with $\varepsilon=64 \cdot \alpha \cdot k^{3} / \Psi$. Since $\left(A_{1}, \ldots, A_{k}\right)$ is an $\alpha$ approximate solution to $\triangle_{k}\left(Y^{\prime}\right)$, we obtain a contradiction

$$
\frac{2 \alpha k^{2}}{\Psi}<\operatorname{Cost}\left(\left\{A_{i}\right\}_{i=1}^{k}\right) \leqslant \alpha \cdot \triangle_{k}\left(Y^{\prime}\right) \leqslant \alpha \cdot \operatorname{Cost}\left(\left\{P_{i}, p^{(i)}\right\}_{i=1}^{k}\right) \leqslant\left(1+\frac{3 k}{\Psi}\right) \cdot \frac{\alpha k^{2}}{\Psi}
$$

Therefore, there exists a permutation $\pi$ (the identity after suitable renumbering of one of the partitions) such that $\mu\left(A_{i} \triangle P_{i}\right)<\frac{8 \alpha \delta}{10^{4} k} \cdot \mu\left(P_{i}\right)$ for all $i \in\{1, \ldots, k\}$.

Part (a.2) of Theorem 5.2 follows from Part (a.1). Indeed, for $\delta^{\prime}=8 \delta / 10^{4}$ we have

$$
\mu\left(A_{i}\right) \geqslant \mu\left(P_{i} \cap A_{i}\right)=\mu\left(P_{i}\right)-\mu\left(P_{i} \backslash A_{i}\right) \geqslant \mu\left(P_{i}\right)-\mu\left(A_{i} \triangle P_{i}\right) \geqslant\left(1-\frac{\alpha \delta^{\prime}}{k}\right) \cdot \mu\left(P_{i}\right)
$$

and $\left|E\left(A_{i}, \overline{A_{i}}\right)\right| \leqslant\left|E\left(P_{i}, \overline{P_{i}}\right)\right|+\mu\left(A_{i} \Delta P_{i}\right)$, since every edge that is counted in $\left|E\left(A_{i}, \overline{A_{i}}\right)\right|$ but not in $\left|E\left(P_{i}, \overline{P_{i}}\right)\right|$ must have an endpoint in $A_{i} \Delta P_{i}$. Thus

$$
\Phi\left(A_{i}\right)=\frac{\left|E\left(A_{i}, \overline{A_{i}}\right)\right|}{\mu\left(A_{i}\right)} \leqslant \frac{\left|E\left(P_{i}, \overline{P_{i}}\right)\right|+\frac{\alpha \delta^{\prime}}{k} \cdot \mu\left(P_{i}\right)}{\left(1-\frac{\alpha \cdot \delta^{\prime}}{k}\right) \cdot \mu\left(P_{i}\right)} \leqslant\left(1+\frac{2 \alpha \delta^{\prime}}{k}\right) \cdot \phi\left(P_{i}\right)+\frac{2 \alpha \delta^{\prime}}{k} .
$$

This completes the proof of Part (a) of Theorem 5.2.

### 6.4 Vectors $\widehat{g_{i}}$ and $f_{i}$ are Close

In this section, we prove Theorem 6.5. We argue in a similar manner as in [PSZ17], but in contrast our results depend on the weaken gap parameter $\Psi$. For completeness, we show in Subsection 6.4.1 that the span of the first $k$ eigenvectors of $\mathcal{L}_{G}$ equals the span of the projections of $P_{i}$ 's characteristic vectors onto the first $k$ eigenvectors. Then, in Subsection 6.4.2, we conclude the proof of Theorem 6.5 by analyzing the eigenvectors $\left\{f_{i}\right\}_{i=1}^{k}$ in terms of projection vectors $\left\{\widehat{f}_{i}\right\}_{i=1}^{k}$.

### 6.4.1 Analyzing the Columns of Matrix $F$

We show now that the span of the first $k$ eigenvectors $\left\{f_{i}\right\}_{i=1}^{k}$ equals the span of the projection vectors $\left\{\widehat{f}_{i}\right\}_{i=1}^{k}$.
Lemma 6.8. If $\Psi>k^{3 / 2}$ then the $\operatorname{span}\left(\left\{\widehat{f}_{i}\right\}_{i=1}^{k}\right)=\operatorname{span}\left(\left\{f_{i}\right\}_{i=1}^{k}\right)$ and thus each eigenvector can be expressed as $f_{i}=\sum_{j=1}^{k} \beta_{j}^{(i)} \cdot \widehat{f}_{j}$ for every $i \in\{1, \ldots, k\}$.

To prove Lemma 6.8, we build upon the following result established by Peng et al. [PSZ17].
Lemma 6.9. [PSZ17, Theorem 1.1 Part 1] For $P_{i} \subset V$ let $\overline{g_{i}}=\frac{D^{1 / 2} \chi_{P_{i}}}{\left\|D^{1 / 2} \chi_{P_{i}}\right\|_{2}}$. Then any $i \in\{1, \ldots, k\}$ it holds that

$$
\left\|\overline{g_{i}}-\widehat{f}_{i}\right\|_{2}^{2}=\sum_{j=k+1}^{n}\left(\alpha_{j}^{(i)}\right)^{2} \leqslant \frac{\mathcal{R}\left(\overline{g_{i}}\right)}{\lambda_{k+1}}=\frac{\phi\left(P_{i}\right)}{\lambda_{k+1}}
$$

Our analysis crucially relies on the following two technical lemmas.
Lemma 6.10. For every $i \in\{1, \ldots, k\}$ and $p \neq q \in\{1, \ldots, k\}$ it holds that

$$
1-\phi\left(P_{i}\right) / \lambda_{k+1} \leqslant\left\|\widehat{f}_{i}\right\|_{2}^{2}=\left\|\alpha^{(i)}\right\|_{2}^{2} \leqslant 1 \quad \text { and } \quad\left|\left\langle\widehat{f}_{p}, \widehat{f}_{q}\right\rangle\right|=\left|\left\langle\alpha^{p}, \alpha^{q}\right\rangle\right| \leqslant \frac{\sqrt{\phi\left(P_{p}\right) \cdot \phi\left(P_{q}\right)}}{\lambda_{k+1}}
$$

Proof. The first part follows by Lemma 6.9 and the following chain of inequalities

$$
1-\frac{\phi\left(P_{i}\right)}{\lambda_{k+1}} \leqslant 1-\sum_{j=k+1}^{n}\left(\alpha_{j}^{(i)}\right)^{2}=\left\|\widehat{f}_{i}\right\|_{2}^{2}=\sum_{j=1}^{k}\left(\alpha_{j}^{(i)}\right)^{2} \leqslant \sum_{j=1}^{n}\left(\alpha_{j}^{(i)}\right)^{2}=1 .
$$

We show now the second part. Since $\left\{f_{i}\right\}_{i=1}^{n}$ are orthonormal eigenvectors we have for all $p \neq q$ that

$$
\begin{equation*}
\left\langle f_{p}, f_{q}\right\rangle=\sum_{l=1}^{n} \alpha_{\ell}^{(p)} \cdot \alpha_{\ell}^{(q)}=0 \tag{6.7}
\end{equation*}
$$

We combine (6.7) and Cauchy-Schwarz to obtain

$$
\begin{aligned}
\left|\left\langle\widehat{f}_{p}, \widehat{f}_{q}\right\rangle\right| & =\left|\sum_{l=1}^{k} \alpha_{\ell}^{(p)} \cdot \alpha_{\ell}^{(q)}\right|=\left|\sum_{l=k+1}^{n} \alpha_{\ell}^{(p)} \cdot \alpha_{\ell}^{(q)}\right| \\
& \leqslant \sqrt{\sum_{l=k+1}^{n}\left(\alpha_{\ell}^{(p)}\right)^{2}} \cdot \sqrt{\sum_{l=k+1}^{n}\left(\alpha_{\ell}^{(q)}\right)^{2}} \leqslant \frac{\sqrt{\phi\left(P_{p}\right) \cdot \phi\left(P_{q}\right)}}{\lambda_{k+1}}
\end{aligned}
$$

Lemma 6.11. If $\Psi>k^{3 / 2}$ then the columns $\left\{F_{:, i}\right\}_{i=1}^{k}$ are linearly independent.
Proof. We show that the columns of matrix $F$ are almost orthonormal. Consider the symmetric matrix $F^{\mathrm{T}} F$. It is known that $\operatorname{ker}\left(F^{\mathrm{T}} F\right)=\operatorname{ker}(F)$ and that all eigenvalues of matrix $F^{\mathrm{T}} F$ are real numbers. We proceeds by showing that the smallest eigenvalue $\lambda_{\text {min }}\left(F^{\mathrm{T}} F\right)>0$. This would imply that $\operatorname{ker}(F)=\emptyset$ and hence yields the statement.

By combining Gersgorin Circle Theorem, Lemma 6.10 and Cauchy-Schwarz it holds that

$$
\begin{aligned}
\lambda_{\min }\left(F^{\mathrm{T}} F\right) & \geqslant \min _{i \in\{1, \ldots, k\}}\left\{\left(F^{\mathrm{T}} F\right)_{i i}-\sum_{j \neq i}^{k}\left|\left(F^{\mathrm{T}} F\right)_{i j}\right|\right\}=\min _{i \in\{1, \ldots, k\}}\left\{\left\|\alpha^{(i)}\right\|_{2}^{2}-\sum_{j \neq i}^{k}\left|\left\langle\alpha^{(j)}, \alpha^{(i)}\right\rangle\right|\right\} \\
& \geqslant 1-\sum_{j=1}^{k} \sqrt{\frac{\phi\left(P_{j}\right)}{\lambda_{k+1}}} \sqrt{\frac{\phi\left(P_{i^{\star}}\right)}{\lambda_{k+1}}} \geqslant 1-\sqrt{k} \sqrt{\sum_{j=1}^{k} \frac{\phi\left(P_{j}\right)}{\lambda_{k+1}}} \sqrt{\frac{\phi\left(P_{i^{\star}}\right)}{\lambda_{k+1}}} \geqslant 1-\frac{k^{3 / 2}}{\Psi}>0
\end{aligned}
$$

where $i^{\star} \in\{1, \ldots, k\}$ is the index that minimizes the expression above.
We present now the proof of Lemma 6.8.
Proof of Lemma 6.8. Let $\nu \in \mathbb{R}^{k}$ be an arbitrary non-zero vector. Notice that

$$
\begin{equation*}
\sum_{i=1}^{k} \nu_{i} \cdot \widehat{f}_{i}=\sum_{i=1}^{k} \nu_{i} \sum_{j=1}^{k} \alpha_{j}^{(i)} f_{j}=\sum_{j=1}^{k}\left(\sum_{i=1}^{k} \nu_{i} \alpha_{j}^{(i)}\right) f_{j}=\sum_{j=1}^{k} \gamma_{j} f_{j}, \quad \text { where } \quad \gamma_{j}=\left\langle F_{j,:}, \nu\right\rangle . \tag{6.8}
\end{equation*}
$$

By Lemma 6.11, the columns $\left\{F_{:, i}\right\}_{i=1}^{k}$ are linearly independent and since $\gamma=F \nu$, it follows that at least one component $\gamma_{j} \neq 0$. Hence, the vectors $\left\{\widehat{f}_{i}\right\}_{i=1}^{k}$ are linearly independent, and since each vector $\widehat{f}_{i}$ is a projection onto the span of the first $k$ eigenvectors $\left\{f_{i}\right\}_{i=1}^{k}$, it follows that $\operatorname{span}\left(\left\{\widehat{f}_{i}\right\}_{i=1}^{k}\right)=\operatorname{span}\left(\left\{f_{i}\right\}_{i=1}^{k}\right)$. Thus, each eigenvector $f_{i}$ can be expressed as a linear combination of the projection vectors $\left\{\widehat{f}_{i}\right\}_{i=1}^{k}$.

### 6.4.2 Analyzing Eigenvectors $f$ in terms of $\widehat{f}_{j}$

In this section, we prove Theorem 6.5. Using Lemma 6.8, we first express each eigenvector $f_{i}=\sum_{j=1}^{k} \beta_{j}^{(i)}$. $\widehat{f}_{j}$ as a linear combination of the projection vectors $\left\{\widehat{f}_{j}\right\}_{j=1}^{k}$, and we bound the squared $\ell_{2}$ norm of the corresponding coefficient vector $\beta^{(i)}=B_{:, i}$ for all $i \in\{1, \ldots, k\}$. Then, we conclude the proof of Theorem 6.5.

Lemma 6.12. If $\Psi>k^{3 / 2}$ then for $i \in[k]$ it holds

$$
\left(1+\frac{2 k}{\Psi}\right)^{-1} \leqslant \sum_{j=1}^{k}\left(\beta_{j}^{(i)}\right)^{2} \leqslant\left(1-\frac{2 k}{\Psi}\right)^{-1}
$$

Proof. We show now the upper bound. By Lemma $6.8 f_{i}=\sum_{j=1}^{k} \beta_{j}^{(i)} \widehat{f}_{j}$ for all $i \in\{1, \ldots, k\}$ and thus

$$
\begin{aligned}
1 & =\left\|f_{i}\right\|_{2}^{2}=\left\langle\sum_{a=1}^{k} \beta_{a}^{(i)} \widehat{f}_{a}, \sum_{b=1}^{k} \beta_{b}^{(i)} \widehat{f}_{b}\right\rangle \\
& =\sum_{j=1}^{k}\left(\beta_{j}^{(i)}\right)^{2}\left\|\widehat{f}_{j}\right\|_{2}^{2}+\sum_{a=1}^{k} \sum_{b \neq a}^{k} \beta_{a}^{(i)} \beta_{b}^{(i)}\left\langle\widehat{f_{a}}, \widehat{f}_{b}\right\rangle \\
& \stackrel{(\star)}{\geqslant}\left(1-\frac{2 k}{\Psi}\right) \cdot \sum_{j=1}^{k}\left(\beta_{j}^{(i)}\right)^{2} .
\end{aligned}
$$

To prove the inequality $(\star)$ we consider the two terms separately.
By Lemma 6.10, $\left\|\widehat{f}_{j}\right\|_{2}^{2} \geqslant 1-\phi\left(P_{j}\right) / \lambda_{k+1}$. We then apply $\sum_{i} a_{i} b_{i} \leqslant\left(\sum_{i} a_{i}\right)\left(\sum_{i} b_{i}\right)$ for all non-negative vectors $a, b$ and obtain

$$
\sum_{j=1}^{k}\left(\beta_{j}^{(i)}\right)^{2}\left(1-\frac{\phi\left(P_{j}\right)}{\lambda_{k+1}}\right)=\sum_{j=1}^{k}\left(\beta_{j}^{(i)}\right)^{2}-\sum_{j=1}^{k}\left(\beta_{j}^{(i)}\right)^{2} \frac{\phi\left(P_{j}\right)}{\lambda_{k+1}} \geqslant\left(1-\frac{k}{\Psi}\right) \sum_{j=1}^{k}\left(\beta_{j}^{(i)}\right)^{2}
$$

Again by Lemma 6.10, we have $\left|\left\langle\widehat{f}_{a}, \widehat{f}_{b}\right\rangle\right| \leqslant \sqrt{\phi\left(P_{a}\right) \phi\left(P_{b}\right)} / \lambda_{k+1}$, and by Cauchy-Schwarz it holds

$$
\begin{aligned}
\sum_{a=1}^{k} \sum_{b \neq a}^{k} \beta_{a}^{(i)} \beta_{b}^{(i)}\left\langle\widehat{f}_{a}, \widehat{f}_{b}\right\rangle & \geqslant-\sum_{a=1}^{k} \sum_{b \neq a}^{k}\left|\beta_{a}^{(i)}\right| \cdot\left|\beta_{b}^{(i)}\right| \cdot\left|\left\langle\widehat{f}_{a}, \widehat{f}_{b}\right\rangle\right| \\
& \geqslant-\frac{1}{\lambda_{k+1}} \sum_{a=1}^{k} \sum_{b \neq a}^{k}\left|\beta_{a}^{(i)}\right| \sqrt{\phi\left(P_{a}\right)} \cdot\left|\beta_{b}^{(i)}\right| \sqrt{\phi\left(P_{b}\right)} \\
& \geqslant-\frac{1}{\lambda_{k+1}}\left(\sum_{j=1}^{k}\left|\beta_{j}^{(i)}\right| \sqrt{\phi\left(P_{j}\right)}\right)^{2} \geqslant-\frac{k}{\Psi} \cdot \sum_{j=1}^{k}\left(\beta_{j}^{(i)}\right)^{2} .
\end{aligned}
$$

The lower bound follows by analogous arguments.
We are now ready to prove Theorem 6.5.
Proof of Theorem 6.5. By Lemma 6.8, we have $f_{i}=\sum_{j=1}^{k} \beta_{j}^{(i)} \widehat{f_{j}}$ and recall that $\widehat{g_{i}}=\sum_{j=1}^{k} \beta_{j}^{(i)} \overline{g_{j}}$ for all $i \in\{1, \ldots, k\}$. Further, by combining triangle inequality, Cauchy-Schwarz, Lemma 6.9 and Lemma 6.12, we obtain that

$$
\begin{aligned}
& \left\|f_{i}-\widehat{g}_{i}\right\|_{2}^{2}=\left\|\sum_{j=1}^{k} \beta_{j}^{(i)}\left(\widehat{f}_{j}-\overline{g_{j}}\right)\right\|_{2}^{2} \leqslant\left(\sum_{j=1}^{k}\left|\beta_{j}^{i}\right| \cdot\left\|\widehat{f}_{j}-\overline{g_{j}}\right\|_{2}\right)^{2} \\
\leqslant & \left(\sum_{j=1}^{k}\left(\beta_{j}^{(i)}\right)^{2}\right) \cdot\left(\sum_{j=1}^{k}\left\|\widehat{f}_{j}-\overline{g_{j}}\right\|_{2}^{2}\right) \leqslant\left(1-\frac{2 k}{\Psi}\right)^{-1}\left(\frac{1}{\lambda_{k+1}} \sum_{j=1}^{k} \phi\left(P_{j}\right)\right) \\
= & \left(1-\frac{2 k}{\Psi}\right)^{-1} \cdot \frac{k}{\Psi} \leqslant\left(1+\frac{3 k}{\Psi}\right) \cdot \frac{k}{\Psi},
\end{aligned}
$$

where the last inequality uses $\Psi>4 k$.

### 6.5 Spectral Properties of Matrix $B$

In this Section, we prove Theorem 6.1 in two steps. In Subsection 6.5.1, we analyzes the column space of matrix $B$ and we show that matrix $B^{\mathrm{T}} B$ is close to the identity matrix. Then, in Subsection 6.5.2, we analyze the row space of matrix $B$ and we prove that matrix $B B^{\mathrm{T}}$ is close to the identity matrix.

### 6.5.1 Analyzing the Column Space of Matrix $B$

We show below that the matrix $B^{\mathrm{T}} B$ is close to the identity matrix.
Lemma 6.13. (Columns) If $\Psi>4 \cdot k^{3 / 2}$ then for all distinct $i, j \in\{1, \ldots, k\}$ it holds

$$
1-\frac{3 k}{\Psi} \leqslant\left\langle B_{:, i}, B_{:, i}\right\rangle \leqslant 1+\frac{3 k}{\Psi} \quad \text { and } \quad\left|\left\langle B_{:, i}, B_{:, j}\right\rangle\right| \leqslant 4 \sqrt{\frac{k}{\Psi}}
$$

Proof. By Lemma 6.12 it holds that

$$
1-\frac{3 k}{\Psi} \leqslant\left\langle B_{:, i}, B_{:, i}\right\rangle=\sum_{j=1}^{k}\left(\beta_{j}^{(i)}\right)^{2} \leqslant 1+\frac{3 k}{\Psi} .
$$

Recall that $\widehat{g_{i}}=\sum_{j=1}^{k} \beta_{j}^{(i)} \cdot \overline{g_{j}}$. Moreover, since the eigenvectors $\left\{f_{i}\right\}_{i=1}^{k}$ and the characteristic vectors $\left\{\bar{g}_{i}\right\}_{i=1}^{k}$ are orthonormal by combing Cauchy-Schwarz and by Theorem 6.5 it holds

$$
\begin{aligned}
\left|\left\langle B_{:, i}, B_{:, j}\right\rangle\right| & =\sum_{l=1}^{k} \beta_{\ell}^{(i)} \beta_{\ell}^{(j)}=\left\langle\sum_{a=1}^{k} \beta_{a}^{(i)} \cdot \overline{g_{a}}, \sum_{b=1}^{k} \beta_{b}^{(j)} \cdot \overline{g_{b}}\right\rangle=\left\langle\widehat{g_{i}}, \widehat{g_{j}}\right\rangle \\
& =\left\langle\left(\widehat{g_{i}}-f_{i}\right)+f_{i},\left(\widehat{g_{j}}-f_{j}\right)+f_{j}\right\rangle \\
& =\left\langle\widehat{g_{i}}-f_{i}, \widehat{g_{j}}-f_{j}\right\rangle+\left\langle\widehat{g_{i}}-f_{i}, f_{j}\right\rangle+\left\langle f_{i}, \widehat{g_{j}}-f_{j}\right\rangle \\
& \leqslant\left\|\widehat{g_{i}}-f_{i}\right\|_{2} \cdot\left\|\widehat{g_{j}}-f_{j}\right\|_{2}+\left\|\widehat{g_{i}}-f_{i}\right\|_{2}+\left\|\widehat{g_{j}}-f_{j}\right\|_{2} \\
& \leqslant\left(1+\frac{3 k}{\Psi}\right) \cdot \frac{k}{\Psi}+2 \sqrt{\left(1+\frac{3 k}{\Psi}\right) \cdot \frac{k}{\Psi}} \leqslant 4 \sqrt{\frac{k}{\Psi}}
\end{aligned}
$$

We demonstrate now that the columns of matrix $B$ are linearly independent.
Lemma 6.14. If $\Psi>25 \cdot k^{3}$ then the columns $\left\{B_{:, i}\right\}_{i=1}^{k}$ are linearly independent.
Proof. Since $\operatorname{ker}(B)=\operatorname{ker}\left(B^{\mathrm{T}} B\right)$ and $B^{\mathrm{T}} B$ is $\operatorname{SPSD}^{1}$ matrix, it suffices to show that the smallest eigenvalue

$$
\lambda\left(B^{\mathrm{T}} B\right)=\min _{x \neq 0} \frac{x^{\mathrm{T}} B^{\mathrm{T}} B x}{x^{\mathrm{T}} x}>0
$$

By Lemma 6.13,

$$
\sum_{i=1}^{k} \sum_{j \neq i}^{k}\left|x_{i}\right|\left|x_{j}\right|\left|\left\langle\beta^{(i)}, \beta^{(j)}\right\rangle\right| \leqslant 4 \sqrt{\frac{k}{\Psi}}\left(\sum_{i=1}^{k}\left|x_{i}\right|\right)^{2} \leqslant\|x\|_{2}^{2} \cdot 4 k \sqrt{\frac{k}{\Psi}}
$$

and

$$
\begin{aligned}
x^{\mathrm{T}} B^{\mathrm{T}} B x & =\left\langle\sum_{i=1}^{k} x_{i} \beta^{(i)}, \sum_{j=1}^{k} x_{j} \beta^{(j)}\right\rangle=\sum_{i=1}^{k} x_{i}^{2}\left\|\beta^{(i)}\right\|_{2}^{2}+\sum_{i=1}^{k} \sum_{j \neq i}^{k} x_{i} x_{j}\left\langle\beta^{(i)}, \beta^{(j)}\right\rangle \\
& \geqslant\left(1-\frac{3 k}{\Psi}\right)\|x\|_{2}^{2}-\sum_{i=1}^{k} \sum_{j \neq i}^{k}\left|x_{i}\right|\left|x_{j}\right|\left|\left\langle\beta^{(i)}, \beta^{(j)}\right\rangle\right| \geqslant\left(1-5 k \sqrt{\frac{k}{\Psi}}\right) \cdot\|x\|_{2}^{2}
\end{aligned}
$$

Hence, $\lambda\left(B^{\mathrm{T}} B\right)>0$ and the statement follows.

### 6.5.2 Analyzing the Row Space of Matrix $B$

In this section, we show that matrix $B B^{\mathrm{T}}$ is close to the identity matrix. We bound now the squared $\ell_{2}$ norm of the rows in matrix $B$, i.e. the diagonal entries in matrix $B B^{\mathrm{T}}$.

Lemma 6.15. (Rows) If $\Psi \geqslant 400 \cdot k^{3} / \varepsilon^{2}$ and $\varepsilon \in(0,1)$ then for all distinct $i, j \in\{1, \ldots, k\}$ it holds

$$
1-\varepsilon \leqslant\left\langle B_{i,:}, B_{i,:}\right\rangle \leqslant 1+\varepsilon
$$

Proof. We show that the eigenvalues of matrix $B B^{\mathrm{T}}$ are concentrated around 1. This would imply that $\chi_{i}^{\mathrm{T}} B B^{\mathrm{T}} \chi_{i}=\left\langle B_{i,:}, B_{i,:}\right\rangle \approx 1$, where $\chi_{i}$ is a characteristic vector. By Lemma 6.13 we have

$$
\left(1-\frac{3 k}{\Psi}\right)^{2} \leqslant\left(\beta^{(i)}\right)^{\mathrm{T}} \cdot B B^{\mathrm{T}} \cdot \beta^{(i)}=\left\|\beta^{(i)}\right\|_{2}^{4}+\sum_{j \neq i}^{k}\left\langle\beta^{(j)}, \beta^{(i)}\right\rangle^{2} \leqslant\left(1+\frac{3 k}{\Psi}\right)^{2}+\frac{16 k^{2}}{\Psi} \leqslant 1+\frac{23 k^{2}}{\Psi}
$$

and

$$
\left|\left(\beta^{(i)}\right)^{\mathrm{T}} \cdot B B^{\mathrm{T}} \cdot \beta^{(j)}\right| \leqslant \sum_{l=1}^{k}\left|\left\langle\beta^{(i)}, \beta^{(l)}\right\rangle\right|\left|\left\langle\beta^{(l)}, \beta^{(j)}\right\rangle\right| \leqslant 8\left(1+\frac{3 k}{\Psi}\right) \sqrt{\frac{k}{\Psi}}+16 \frac{k^{2}}{\Psi} \leqslant 11 \sqrt{\frac{k}{\Psi}}
$$

[^9]By Lemma 6.14 every vector $x \in \mathbb{R}^{k}$ can be expressed as $x=\sum_{i=1}^{k} \gamma_{i} \beta^{(i)}$.

$$
\begin{aligned}
x^{\mathrm{T}} B B^{\mathrm{T}} x & =\sum_{i=1}^{k} \gamma_{i}\left(\beta^{(i)}\right)^{\mathrm{T}} \cdot B B^{\mathrm{T}} \cdot \sum_{j=1}^{k} \gamma_{j} \beta^{(j)} \\
& =\sum_{i=1}^{k} \gamma_{i}^{2}\left(\beta^{(i)}\right)^{\mathrm{T}} \cdot B B^{\mathrm{T}} \cdot \beta^{(i)}+\sum_{i=1}^{k} \sum_{j \neq i}^{k} \gamma_{i} \gamma_{j}\left(\beta^{(i)}\right)^{\mathrm{T}} \cdot B B^{\mathrm{T}} \cdot \beta^{(j)} \\
& \geqslant\left(1-\frac{23 k^{2}}{\Psi}-11 \cdot k \sqrt{\frac{k}{\Psi}}\right)\|\gamma\|_{2}^{2} \geqslant\left(1-14 \cdot k \sqrt{\frac{k}{\Psi}}\right)\|\gamma\|_{2}^{2}
\end{aligned}
$$

and

$$
x^{\mathrm{T}} x=\sum_{i=1}^{k} \sum_{j=1}^{k} \gamma_{i} \gamma_{j}\left\langle\beta^{(i)}, \beta^{(j)}\right\rangle=\sum_{i=1}^{k} \gamma_{i}^{2}\left\|\beta^{(i)}\right\|_{2}^{2}+\sum_{i=1}^{k} \sum_{j \neq i}^{k} \gamma_{i} \gamma_{j}\left\langle\beta^{(i)}, \beta^{(j)}\right\rangle
$$

By Lemma 6.13 we have $\left|\sum_{i=1}^{k} \sum_{j \neq i}^{k} \gamma_{i} \gamma_{j}\left\langle\beta^{(i)}, \beta^{(j)}\right\rangle\right| \leqslant\|\gamma\|_{2}^{2} \cdot 4 k \sqrt{\frac{k}{\Psi}}$ and $\left\|\beta^{(i)}\right\|_{2}^{2} \leqslant 1+\frac{3 k}{\Psi}$. Thus, it holds

$$
\left(1-5 k \sqrt{\frac{k}{\Psi}}\right)\|\gamma\|_{2}^{2} \leqslant x^{\mathrm{T}} x \leqslant\left(1+5 k \sqrt{\frac{k}{\Psi}}\right)\|\gamma\|_{2}^{2}
$$

Hence, we have

$$
1-20 k \sqrt{\frac{k}{\Psi}} \leqslant \lambda\left(B B^{\mathrm{T}}\right) \leqslant 1+20 k \sqrt{\frac{k}{\Psi}}
$$

This proves the first part of Theorem 6.1. We turn now to the second part and restate it in the following Lemma.

Lemma 6.16. (Rows) If $\Psi \geqslant 10^{4} \cdot k^{3} / \varepsilon^{2}$ and $\varepsilon \in(0,1)$ then for all distinct $i, j \in\{1, \ldots, k\}$ it holds

$$
\mid\left\langle B_{i,:}, B_{j,:}\right| \mid \leqslant \sqrt{\varepsilon} .
$$

Let $E \in \mathbb{R}^{k \times k}$ be a symmetric matrix such that $B^{\mathrm{T}} B=I+E$ and $\left|E_{i j}\right| \leqslant 4 \sqrt{k / \Psi}$. Then,

$$
\begin{equation*}
\left(B B^{\mathrm{T}}\right)^{2}=B(I+E) B^{\mathrm{T}}=B B^{\mathrm{T}}+B E B^{\mathrm{T}} \tag{6.9}
\end{equation*}
$$

We show next that the absolute value of every eigenvalue of matrix $B E B^{\mathrm{T}}$ is small, and further demonstrate that this implies that all entries of matrix $B E B^{\mathrm{T}}$ are small. Then, we conclude the proof of Lemma 6.16.

Lemma 6.17. If $\Psi \geqslant 40^{2} \cdot k^{3} / \varepsilon^{2}$ and $\varepsilon \in(0,1)$, then every eigenvalue $\lambda$ of matrix $B E B^{T}$ satisfies

$$
\left|\lambda\left(B E B^{\mathrm{T}}\right)\right| \leqslant \varepsilon / 5
$$

Proof. Let $z=B^{\mathrm{T}} x$. We upper bound the quadratic form

$$
\left|x^{\mathrm{T}} B E B^{\mathrm{T}} x\right|=\left|z^{\mathrm{T}} E z\right| \leqslant \sum_{i j}\left|E_{i j}\right|\left|z_{i}\right|\left|z_{j}\right| \leqslant 4 \sqrt{\frac{k}{\Psi}} \cdot\left(\sum_{i=1}^{k}\left|z_{i}\right|\right)^{2} \leqslant\|z\|_{2}^{2} \cdot 4 k \sqrt{\frac{k}{\Psi}} .
$$

By Lemma 6.15, we have $1-\varepsilon \leqslant \lambda\left(B B^{\mathrm{T}}\right) \leqslant 1+\varepsilon$ and since $\|z\|_{2}^{2}=\frac{x B B^{\mathrm{T}} x}{x^{\mathrm{T}} x} \cdot\|x\|_{2}^{2}$, it follows that

$$
\frac{\|z\|_{2}^{2}}{1+\varepsilon} \leqslant\|x\|_{2}^{2} \leqslant \frac{\|z\|_{2}^{2}}{1-\varepsilon}
$$

and hence

$$
\left|\lambda\left(B E B^{\mathrm{T}}\right)\right| \leqslant \max _{x} \frac{\left|x^{\mathrm{T}} B E B^{\mathrm{T}} x\right|}{x^{\mathrm{T}} x} \leqslant 4(1+\varepsilon) \cdot k \sqrt{\frac{k}{\Psi}} \leqslant \varepsilon / 5 .
$$

Lemma 6.18. If $\Psi \geqslant 40^{2} \cdot k^{3} / \varepsilon^{2}$ and $\varepsilon \in(0,1)$, then it holds that $\left|\left(B E B^{T}\right)_{i j}\right| \leqslant \varepsilon / 5$ for every $i, j \in$ $\{1, \ldots, k\}$.

Proof. Since matrix $E \in \mathbb{R}^{k \times k}$ is a symmetric, by construction matrix, $B E B^{\mathrm{T}} \in \mathbb{R}^{k \times k}$ is also symmetric. Using the SVD Theorem, there is an orthonormal basis $\left\{u_{i}\right\}_{i=1}^{k}$ such that $B E B^{\mathrm{T}}=\sum_{i=1}^{k} \lambda_{i}\left(B E B^{\mathrm{T}}\right) \cdot u_{i} u_{i}^{\mathrm{T}}$. Thus, it suffices to bound the expression

$$
\left|\left(B E B^{\mathrm{T}}\right)_{i j}\right| \leqslant \sum_{l=1}^{k}\left|\lambda_{\ell}\left(B E B^{\mathrm{T}}\right)\right| \cdot\left|\left(u_{\ell} u_{\ell}^{\mathrm{T}}\right)_{i j}\right| .
$$

Let $U \in \mathbb{R}^{k \times k}$ be a square matrix whose $i$-th column is vector $u_{i}$. By construction, matrix $U$ is orthogonal and satisfies $U^{\mathrm{T}} U=I=U U^{\mathrm{T}}$. In particular, it holds that $\left\|U_{i,}:\right\|_{2}^{2}=1$, for all $i$. Therefore, we have

$$
\sum_{l=1}^{k}\left|\left(u_{\ell}\right)_{i}\right| \cdot\left|\left(u_{\ell}\right)_{j}\right| \leqslant \sqrt{\left\|U_{i,:}\right\|_{2}^{2}} \sqrt{\left\|U_{j,:}\right\|_{2}^{2}}=1
$$

We apply now Lemma 6.17 to obtain

$$
\sum_{l=1}^{k}\left|\lambda_{\ell}\left(B E B^{\mathrm{T}}\right)\right| \cdot\left|\left(u_{\ell} u_{\ell}^{\mathrm{T}}\right)_{i j}\right| \leqslant \frac{\varepsilon}{5} \cdot \sum_{l=1}^{k}\left|\left(u_{\ell}\right)_{i}\right| \cdot\left|\left(u_{\ell}\right)_{j}\right| \leqslant \frac{\varepsilon}{5}
$$

We are now ready to prove Lemma 6.16.
Proof of Lemma 6.16. By (6.9) we have $\left(B B^{\mathrm{T}}\right)^{2}=B B^{\mathrm{T}}+B E B^{\mathrm{T}}$. Observe that the $(i, j)$-th entry of matrix $B B^{\mathrm{T}}$ is equal to the inner product between the $i$-th and $j$-th row of matrix $B$, i.e. $\left(B B^{\mathrm{T}}\right)_{i j}=$ $\left\langle B_{i,:}, B_{j,:}\right\rangle$. Moreover, we have

$$
\left[\left(B B^{\mathrm{T}}\right)^{2}\right]_{i j}=\sum_{l=1}^{k}\left(B B^{\mathrm{T}}\right)_{i, l}\left(B B^{\mathrm{T}}\right)_{l, j}=\sum_{l=1}^{k}\left\langle B_{i,:}, B_{l,:}\right\rangle\left\langle B_{l,:}, B_{j,:\rangle}\right\rangle
$$

For the entries on the main diagonal, it holds

$$
\left\langle B_{i,:}, B_{i,:}\right\rangle^{2}+\sum_{l \neq i}^{k}\left\langle B_{i,:}, B_{l,:}\right\rangle^{2}=\left[\left(B B^{\mathrm{T}}\right)^{2}\right]_{i i}=\left[B B^{\mathrm{T}}+B E B^{\mathrm{T}}\right]_{i i}=\left\langle B_{i,:}, B_{i,:}\right\rangle+\left(B E B^{\mathrm{T}}\right)_{i i},
$$

and hence by applying Lemma 6.15 with $\varepsilon^{\prime}=\varepsilon / 5$ and Lemma 6.18 with $\varepsilon^{\prime}=\varepsilon$ we obtain

$$
\left\langle B_{i,:}, B_{j,:}\right\rangle^{2} \leqslant \sum_{l \neq i}\left\langle B_{i,:}, B_{l,:}\right\rangle^{2} \leqslant\left(1+\frac{\varepsilon}{5}\right)+\frac{\varepsilon}{5}-\left(1-\frac{\varepsilon}{5}\right)^{2} \leqslant \varepsilon .
$$

### 6.6 Volume Overlap Lemma

In this section, we prove Lemma 6.4. Our main technical contribution is to strengthen the lower bound of $k$-means cost in [PSZ17, Lemma 4.5] by a factor of $k$, under the weaken gap assumption.

We begin by stating a useful Corollary of Lemma 6.3.
Corollary 6.19. Let $\Psi=20^{4} \cdot k^{3} / \delta$ for some $\delta \in(0,1 / 2]$. Suppose $c_{i}$ is the center of a cluster $A_{i}$. If $\left\|c_{i}-p^{\left(i_{1}\right)}\right\|_{2} \geqslant\left\|c_{i}-p^{\left(i_{2}\right)}\right\|_{2}$, then it holds that

$$
\left\|c_{i}-p^{\left(i_{1}\right)}\right\|_{2}^{2} \geqslant \frac{1}{4}\left\|p^{\left(i_{1}\right)}-p^{\left(i_{2}\right)}\right\|_{2}^{2} \geqslant\left[8 \cdot \min \left\{\mu\left(P_{i_{1}}\right), \mu\left(P_{i_{2}}\right)\right\}\right]^{-1} .
$$

We restate now [PSZ17, Lemma 4.6] whose analysis crucially relies on the following function

$$
\begin{equation*}
\sigma(\ell)=\arg \max _{j \in\{1, \ldots, k\}} \frac{\mu\left(A_{\ell} \cap P_{j}\right)}{\mu\left(P_{j}\right)}, \quad \text { for all } \ell \in\{1, \ldots, k\} \tag{6.10}
\end{equation*}
$$

Lemma 6.20. [PSZ17, Lemma 4.6] Let $\left(P_{1}, \ldots, P_{k}\right)$ and $\left(A_{1}, \ldots, A_{k}\right)$ be $k$-way node partitions of $G$. Suppose for every permutation $\pi:\{1, \ldots, k\} \rightarrow\{1, \ldots, k\}$ there is an index $i \in\{1, \ldots, k\}$ such that

$$
\begin{equation*}
\mu\left(A_{i} \triangle P_{\pi(i)}\right) \geqslant 2 \varepsilon \cdot \mu\left(P_{\pi(i)}\right), \tag{6.11}
\end{equation*}
$$

where $\varepsilon \in(0,1 / 2)$ is a parameter. Then one of the following three statements holds:

1. If $\sigma$ is a permutation and $\mu\left(P_{\sigma(i)} \backslash A_{i}\right) \geqslant \varepsilon \cdot \mu\left(P_{\sigma(i)}\right)$, then for every index $j \neq i$ there is a real $\varepsilon_{j} \geqslant 0$ such that

$$
\mu\left(A_{j} \cap P_{\sigma(j)}\right) \geqslant \mu\left(A_{j} \cap P_{\sigma(i)}\right) \geqslant \varepsilon_{j} \cdot \min \left\{\mu\left(P_{\sigma(j)}\right), \mu\left(P_{\sigma(i)}\right)\right\},
$$

and $\sum_{j \neq i} \varepsilon_{j} \geqslant \varepsilon$.
2. If $\sigma$ is a permutation and $\mu\left(A_{i} \backslash P_{\sigma(i)}\right) \geqslant \varepsilon \cdot \mu\left(P_{\sigma(i)}\right)$, then for every $j \neq i$ there is a real $\varepsilon_{j} \geqslant 0$ such that

$$
\mu\left(A_{i} \cap P_{\sigma(i)}\right) \geqslant \varepsilon_{j} \cdot \mu\left(P_{\sigma(i)}\right), \quad \mu\left(A_{i} \cap P_{\sigma(j)}\right) \geqslant \varepsilon_{j} \cdot \mu\left(P_{\sigma(i)}\right)
$$

and $\sum_{j \neq i} \varepsilon_{j} \geqslant \varepsilon$.
3. If $\sigma$ is not a permutation, then there is an index $\ell \notin\{\sigma(1), \ldots, \sigma(k)\}$ and for every index $j$ there is a real $\varepsilon_{j} \geqslant 0$ such that

$$
\mu\left(A_{j} \cap P_{\sigma(j)}\right) \geqslant \mu\left(A_{j} \cap P_{\ell}\right) \geqslant \varepsilon_{j} \cdot \min \left\{\mu\left(P_{\sigma(j)}\right), \mu\left(P_{\ell}\right)\right\}
$$

and $\sum_{j=1}^{k} \varepsilon_{j}=1$.
We strengthen now the lower bound of $k$-means cost in [PSZ17, Lemma 4.5] by a factor of $k$.
Lemma 6.21. Suppose the hypothesis of Lemma 6.20 is satisfied and $\Psi=20^{4} \cdot k^{3} / \delta$ for some $\delta \in(0,1 / 2]$. Then it holds

$$
\operatorname{Cost}\left(\left\{A_{i}, c_{i}\right\}_{i=1}^{k}\right) \geqslant \frac{\varepsilon}{16}-\frac{2 k^{2}}{\Psi}
$$

Proof. By definition

$$
\begin{equation*}
\operatorname{Cost}\left(\left\{A_{i}, c_{i}\right\}_{i=1}^{k}\right)=\sum_{i=1}^{k} \sum_{j=1}^{k} \sum_{u \in A_{i} \cap P_{j}} d(u)\left\|\mathcal{F}(u)-c_{i}\right\|_{2}^{2} \stackrel{\text { def }}{=} \Lambda . \tag{6.12}
\end{equation*}
$$

Since for every vectors $x, y, z \in \mathbb{R}^{k}$ it holds

$$
2\left(\|x-y\|_{2}^{2}+\|z-y\|_{2}^{2}\right) \geqslant\left(\|x-y\|_{2}+\|z-y\|_{2}\right)^{2} \geqslant\|x-z\|_{2}^{2}
$$

we have for all indices $i, j \in\{1, \ldots, k\}$ that

$$
\begin{equation*}
\left\|\mathcal{F}(u)-c_{i}\right\|_{2}^{2} \geqslant \frac{\left\|p^{(j)}-c_{i}\right\|_{2}^{2}}{2}-\left\|\mathcal{F}(u)-p^{(j)}\right\|_{2}^{2} \tag{6.13}
\end{equation*}
$$

Our proof proceeds by considering three cases. Let $i \in\{1, \ldots, k\}$ be the index from the hypothesis in Lemma 6.20.

Case 1. Suppose the first conclusion of Lemma 6.20 holds. For every index $j \neq i$ let

$$
p^{\gamma(j)}= \begin{cases}p^{\sigma(j)} & , \text { if }\left\|p^{\sigma(j)}-c_{j}\right\|_{2} \geqslant\left\|p^{\sigma(i)}-c_{j}\right\|_{2} \\ p^{\sigma(i)} & , \text { otherwise }\end{cases}
$$

Then by combining (6.13), Corollary 6.19 and Lemma 6.6 , we have

$$
\begin{aligned}
\Lambda & \geqslant \frac{1}{2} \sum_{j \neq i} \sum_{u \in A_{j} \cap P_{\gamma(j)}} d(u)\left\|p^{\gamma(j)}-c_{j}\right\|_{2}^{2}-\sum_{j \neq i} \sum_{u \in A_{j} \cap P_{\gamma(j)}}\left\|\mathcal{F}(u)-p^{\gamma(j)}\right\|_{2}^{2} \\
& \geqslant \frac{1}{16} \sum_{j \neq i} \frac{\mu\left(A_{j} \cap P_{\gamma(j)}\right)}{\min \left\{\mu\left(P_{\sigma(i)}\right), \mu\left(P_{\sigma(j)}\right)\right\}}-\left(1+\frac{3 k}{\Psi}\right) \cdot \frac{k^{2}}{\Psi} \geqslant \frac{\varepsilon}{16}-\frac{2 k^{2}}{\Psi} .
\end{aligned}
$$

Case 2. Suppose the second conclusion of Lemma 6.20 holds. Notice that if $\mu\left(A_{i} \cap P_{\sigma(i)}\right) \leqslant(1-\varepsilon)$. $\mu\left(P_{\sigma(i)}\right)$ then $\mu\left(P_{\sigma(i)} \backslash A_{i}\right) \geqslant \varepsilon \cdot \mu\left(P_{\sigma(i)}\right)$ and thus we can argue as in Case 1 . Hence, we can assume that it holds

$$
\begin{equation*}
\mu\left(A_{i} \cap P_{\sigma(i)}\right) \geqslant(1-\varepsilon) \cdot \mu\left(P_{\sigma(i)}\right) \tag{6.14}
\end{equation*}
$$

We proceed by analyzing two subcases.
a) If $\left\|p^{\sigma(j)}-c_{i}\right\|_{2} \geqslant\left\|p^{\sigma(i)}-c_{i}\right\|$ holds for all $j \neq i$ then by combining (6.13), Corollary 6.19 and Lemma 6.6 it follows

$$
\begin{aligned}
\Lambda & \geqslant \frac{1}{2} \sum_{j \neq i} \sum_{u \in A_{i} \cap P_{\sigma(j)}} d(u)\left\|p^{\sigma(j)}-c_{i}\right\|_{2}^{2}-\sum_{j \neq i} \sum_{u \in A_{i} \cap P_{\sigma(j)}}\left\|\mathcal{F}(u)-p^{\sigma(j)}\right\|_{2}^{2} \\
& \geqslant \frac{1}{2} \sum_{j \neq i} \frac{\mu\left(A_{i} \cap P_{\sigma(j)}\right)}{\min \left\{\mu\left(P_{\sigma(i)}\right), \mu\left(P_{\sigma(j)}\right)\right\}}-\left(1+\frac{3 k}{\Psi}\right) \cdot \frac{k^{2}}{\Psi} \geqslant \frac{\varepsilon}{16}-\frac{2 k^{2}}{\Psi} .
\end{aligned}
$$

b) Suppose there is an index $j \neq i$ such that $\left\|p^{\sigma(j)}-c_{i}\right\|_{2}<\left\|p^{\sigma(i)}-c_{i}\right\|$. Then by triangle inequality combined with Corollary 6.19 we have

$$
\left\|p^{\sigma(i)}-c_{i}\right\|_{2}^{2} \geqslant \frac{1}{4}\left\|p^{\sigma(i)}-p^{\sigma(j)}\right\|_{2} \geqslant\left[8 \cdot \min \left\{\mu\left(P_{\sigma(i)}\right), \mu\left(P_{\sigma(j)}\right)\right\}\right]^{-1}
$$

Thus, by combining (6.13), (6.14) and Lemma 6.6 we obtain

$$
\begin{aligned}
\Lambda & \geqslant \frac{1}{2} \sum_{u \in A_{i} \cap P_{\sigma(i)}} d(u)\left\|p^{\sigma(i)}-c_{i}\right\|_{2}^{2}-\sum_{u \in A_{i} \cap P_{\sigma(i)}} d(u)\left\|\mathcal{F}(u)-p^{\sigma(i)}\right\|_{2}^{2} \\
& \geqslant \frac{1}{16} \cdot \frac{\mu\left(A_{i} \cap P_{\sigma(i)}\right)}{\min \left\{\mu\left(P_{\sigma(i)}\right), \mu\left(P_{\sigma(j)}\right)\right\}}-\left(1+\frac{3 k}{\Psi}\right) \cdot \frac{k^{2}}{\Psi} \geqslant \frac{1-\varepsilon}{16}-\frac{2 k^{2}}{\Psi} .
\end{aligned}
$$

Case 3. Suppose the third conclusion of Lemma 6.20 holds, i.e., $\sigma$ is not a permutation. Then there is an index $\ell \in\{1, \ldots, k\} \backslash\{\sigma(1), \ldots, \sigma(k)\}$ and for every index $j \in\{1, \ldots, k\}$ let

$$
p^{\gamma(j)}= \begin{cases}p^{\ell} & , \text { if }\left\|p^{\ell}-c_{j}\right\|_{2} \geqslant\left\|p^{\sigma(j)}-c_{j}\right\|_{2} \\ p^{\sigma(j)} & , \text { otherwise }\end{cases}
$$

By combining (6.13), Corollary 6.19 and Lemma 6.6 it follows that

$$
\begin{aligned}
\Lambda & \geqslant \frac{1}{2} \sum_{j=1}^{k} \sum_{u \in A_{j} \cap P_{\gamma(j)}} d(u)\left\|p^{\gamma(j)}-c_{j}\right\|_{2}^{2}-\sum_{j=1}^{k} \sum_{u \in A_{j} \cap P_{\gamma(j)}} d(u)\left\|\mathcal{F}(u)-p^{\gamma(j)}\right\|_{2}^{2} \\
& \geqslant \frac{1}{16} \sum_{j=1}^{k} \frac{\mu\left(A_{j} \cap P_{\gamma(j)}\right)}{\min \left\{\mu\left(P_{\sigma(j)}\right), \mu\left(P_{\ell}\right)\right\}}-\left(1+\frac{3 k}{\Psi}\right) \cdot \frac{k^{2}}{\Psi} \geqslant \frac{1}{16}-\frac{2 k^{2}}{\Psi} .
\end{aligned}
$$

We are now ready to prove Lemma 6.4.
Proof of Lemma 6.4. We apply Lemma 6.20 with $\varepsilon^{\prime}=\varepsilon / k$. Then, by Lemma 6.21 we have

$$
\operatorname{Cost}\left(\left\{A_{i}, c_{i}\right\}_{i=1}^{k}\right) \geqslant \frac{\varepsilon}{16 k}-\frac{2 k^{2}}{\Psi},
$$

and the desired result follows by setting $\varepsilon \geqslant 64 \alpha \cdot k^{3} / \Psi$.

## Chapter 7

## Analysis of Approximate Spectral Clustering

### 7.1 Normalized Spectral Embedding

In this section, we prove Theorem 5.4, showing that the normalized $\mathrm{SE} Y^{\prime}$ is $\varepsilon$-separated. For convenience of the reader, we restate the result.

Theorem 5.4 (from page 62). Let $G$ be a graph that satisfies $\Psi=20^{4} \cdot k^{3} / \delta, \delta \in(0,1 / 2]$ and $k / \delta \geqslant 10^{9}$. Then for $\varepsilon=6 \cdot 10^{-7}$ it holds

$$
\begin{equation*}
\triangle_{k}\left(Y^{\prime}\right) \leqslant \varepsilon^{2} \cdot \triangle_{k-1}\left(Y^{\prime}\right) \tag{7.1}
\end{equation*}
$$

We establish first a lower bound on $\triangle_{k-1}\left(Y^{\prime}\right)$.
Lemma 7.1. Let $G$ be a graph that satisfies $\Psi=20^{4} \cdot k^{3} / \delta$ for some $\delta \in(0,1 / 2]$. Then for $\delta^{\prime}=2 \delta / 20^{4}$ it holds

$$
\begin{equation*}
\triangle_{k-1}\left(Y^{\prime}\right) \geqslant \frac{1}{12}-\frac{\delta^{\prime}}{k} \tag{7.2}
\end{equation*}
$$

Before we present the proof of Lemma 7.1, we show that it implies (7.1). By Lemma 6.6, we have

$$
\triangle_{k}\left(Y^{\prime}\right) \leqslant \frac{2 k^{2}}{\Psi}=\frac{\delta^{\prime}}{k}
$$

and thus, by applying Lemma 7.1 with $k / \delta \geqslant 10^{9}$ and $\varepsilon=6 \cdot 10^{-7}$, we obtain

$$
\triangle_{k-1}\left(Y^{\prime}\right) \geqslant \frac{1}{12}-\frac{\delta^{\prime}}{k}=\frac{1}{12}-\frac{2}{20^{4}} \cdot \frac{\delta}{k} \geqslant \frac{10^{10}}{9 \cdot 2^{5}} \cdot \frac{\delta}{k}=\frac{1}{\varepsilon^{2}} \cdot \frac{\delta^{\prime}}{k} \geqslant \frac{1}{\varepsilon^{2}} \cdot \triangle_{k}\left(Y^{\prime}\right)
$$

### 7.1.1 Proof of Lemma 7.1

We argue in a similar manner as in Lemma 6.21 (c.f. Case 3). We start by giving some notation, then we establish Lemma 7.2 and apply it in the proof of Lemma 7.1.

We redefine the function $\sigma$, see (6.10), such that for any two partitions $\left(P_{1}, \ldots, P_{k}\right)$ and $\left(Z_{1}, \ldots, Z_{k-1}\right)$ of $V$, we define a mapping $\sigma:\{1, \ldots, k-1\} \mapsto\{1, \ldots, k\}$ by

$$
\sigma(i)=\arg \max _{j \in\{1, \ldots, k\}} \frac{\mu\left(Z_{i} \cap P_{j}\right)}{\mu\left(P_{j}\right)}, \quad \text { for every } i \in\{1, \ldots, k-1\} .
$$

We lower bound now the overlapping of clusters between any $k$-way and ( $k-1$ )-way partitions of $V$ in terms of volume.

Lemma 7.2. Suppose $\left(P_{1}, \ldots, P_{k}\right)$ and $\left(Z_{1}, \ldots, Z_{k-1}\right)$ are partitions of $V$. Then for any index $\ell \in$ $\{1, \ldots, k\} \backslash\{\sigma(1), \ldots, \sigma(k-1)\}$ (there is at least one such $\ell$ ) and for every $i \in\{1, \ldots, k-1\}$ it holds

$$
\left\{\mu\left(Z_{i} \cap P_{\sigma(i)}\right), \mu\left(Z_{i} \cap P_{\ell}\right)\right\} \geqslant \tau_{i} \cdot \min \left\{\mu\left(P_{\ell}\right), \mu\left(P_{\sigma(i)}\right)\right\},
$$

where $\sum_{i=1}^{k-1} \tau_{i}=1$ and $\tau_{i} \geqslant 0$.
Proof. By pigeonhole principle there is an index $\ell \in\{1, \ldots, k\}$ such that $\ell \notin\{\sigma(1), \ldots, \sigma(k-1)\}$. Thus, for every $i \in\{1, \ldots, k-1\}$ we have $\sigma(i) \neq \ell$ and

$$
\frac{\mu\left(Z_{i} \cap P_{\sigma(i)}\right)}{\mu\left(P_{\sigma(i)}\right)} \geqslant \frac{\mu\left(Z_{i} \cap P_{\ell}\right)}{\mu\left(P_{\ell}\right)} \stackrel{\text { def }}{=} \tau_{i},
$$

where $\sum_{i=1}^{k-1} \tau_{i}=1$ and $\tau_{i} \geqslant 0$ for all $i$. Hence, the statement follows.

We present now the proof of Lemma 7.1.

Proof of Lemma 7.1. Let $\left(Z_{1}, \ldots, Z_{k-1}\right)$ be a $(k-1)$-way partition of $V$ with centers $c_{1}^{\prime}, \ldots, c_{k-1}^{\prime}$ that achieves $\triangle_{k-1}\left(Y^{\prime}\right)$, and $\left(P_{1}, \ldots, P_{k}\right)$ be a $k$-way partition of $V$ achieving $\widehat{\rho}_{\text {avr }}(k)$. Our goal now is to lower bound the optimum $(k-1)$-means cost

$$
\begin{equation*}
\triangle_{k-1}\left(Y^{\prime}\right)=\sum_{i=1}^{k-1} \sum_{j=1}^{k} \sum_{u \in Z_{i} \cap P_{j}} d_{u}\left\|\mathcal{F}(u)-c_{i}^{\prime}\right\|_{2}^{2} \tag{7.3}
\end{equation*}
$$

By Lemma 7.2 there is an index $\ell \in\{1, \ldots, k\} \backslash\{\sigma(1), \ldots, \sigma(k-1)\}$. For $i \in\{1, \ldots, k-1\}$ let

$$
p^{\gamma(i)}= \begin{cases}p^{\ell} & , \text { if }\left\|p^{\ell}-c_{i}^{\prime}\right\|_{2} \geqslant\left\|p^{\sigma(i)}-c_{i}^{\prime}\right\|_{2} \\ p^{\sigma(i)} & , \text { otherwise }\end{cases}
$$

Then by combining Corollary 6.19 and Lemma 7.2, we have

$$
\begin{equation*}
\left\|p^{\gamma(i)}-c_{i}^{\prime}\right\|_{2}^{2} \geqslant\left[8 \cdot \min \left\{\mu\left(P_{\ell}\right), \mu\left(P_{\sigma(i)}\right)\right\}\right]^{-1} \text { and } \mu\left(Z_{i} \cap P_{\gamma(i)}\right) \geqslant \tau_{i} \cdot \min \left\{\mu\left(P_{\ell}\right), \mu\left(P_{\sigma(i)}\right)\right\}, \tag{7.4}
\end{equation*}
$$

where $\sum_{i=1}^{k-1} \tau_{i}=1$. We now lower bound the expression in (7.3). Since

$$
\left\|\mathcal{F}(u)-c_{i}^{\prime}\right\|_{2}^{2} \geqslant \frac{1}{2}\left\|p^{\gamma(i)}-c_{i}^{\prime}\right\|_{2}^{2}-\left\|\mathcal{F}(u)-p^{\gamma(i)}\right\|_{2}^{2},
$$

it follows for $\delta^{\prime}=2 \delta / 20^{4}$ that

$$
\begin{aligned}
\triangle_{k-1}\left(\mathcal{X}_{V}\right) & =\sum_{i=1}^{k-1} \sum_{j=1}^{k} \sum_{u \in Z_{i} \cap P_{j}} d_{u}\left\|\mathcal{F}(u)-c_{i}^{\prime}\right\|_{2}^{2} \geqslant \sum_{i=1}^{k-1} \sum_{u \in Z_{i} \cap P_{\gamma(i)}} d_{u}\left\|\mathcal{F}(u)-c_{i}^{\prime}\right\|_{2}^{2} \\
& \geqslant \frac{1}{2} \sum_{i=1}^{k-1} \sum_{u \in Z_{i} \cap P_{\gamma(i)}} d_{u}\left\|p^{\gamma(i)}-c_{i}^{\prime}\right\|_{2}^{2}-\sum_{i=1}^{k-1} \sum_{u \in Z_{i} \cap P_{\gamma(i)}} d_{u}\left\|\mathcal{F}(u)-p^{\gamma(i)}\right\|_{2}^{2} \\
& \geqslant \frac{1}{2} \sum_{i=1}^{k-1} \frac{\mu\left(Z_{i} \cap P_{\gamma(i)}\right)}{8 \cdot \min \left\{\mu\left(P_{\gamma(i)}\right), \mu\left(P_{\sigma(i)}\right)\right\}}-\sum_{i=1}^{k} \sum_{u \in P_{i}} d_{u}\left\|\mathcal{F}(u)-p^{i}\right\|_{2}^{2} \\
& \geqslant \frac{1}{16}-\frac{\delta^{\prime}}{k}
\end{aligned}
$$

where the last inequality holds due to (7.4) and Lemma 6.6.

### 7.2 Approximate Normalized Spectral Embedding

In this section, we prove Theorem 5.6 , which shows that the approximate normalized $\mathrm{SE} \widetilde{Y^{\prime}}$, computed via the Power method, is $\varepsilon$-separated.

Before we state our results, we need some notation. Let $X_{\mathrm{opt}}^{\prime}$ be an indicator matrix, see (5.11), corresponding to an optimal $k$-way row partition of the normalized SE $Y^{\prime}$. Then, the optimum $k$-means cost of $Y^{\prime}$ in matrix notation reads

$$
\triangle_{k}\left(Y^{\prime}\right)=\left\|Y^{\prime}-X_{\mathrm{opt}}^{\prime}\left(X_{\mathrm{opt}}^{\prime}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F}^{2}
$$

Similarly, for the approximate normalized SE $\widetilde{Y^{\prime}}$, let $\widetilde{X_{\mathrm{opt}}^{\prime}}$ be an indicator matrix such that

$$
\triangle_{k}\left(\widetilde{Y^{\prime}}\right)=\left\|\widetilde{Y^{\prime}}-\widetilde{X_{\mathrm{opt}}^{\prime}}\left(\widetilde{X_{\mathrm{opt}}^{\prime}}\right)^{\mathrm{T}} \widetilde{Y^{\prime}}\right\|_{F}^{2}
$$

In Subsection 7.2.1, using techniques from [BKG15, Lemma 5] and [BM14, Lemma 7], we prove the following statement.

Lemma 7.3. Let $\lambda_{k}$ and $\lambda_{k+1}$ be the $k$-th and $(k+1)$-st smallest eigenvalue of $\mathcal{L}_{G}, Y$ be the canonical $S E$, and $S \in \mathbb{R}^{n \times k}$ be a matrix whose entries are i.i.d. samples from the standard Gaussian distribution. For any $\beta, \varepsilon \in(0,1)$ and $p \geqslant \ln (8 n k / \varepsilon \beta) / \ln \left(1 / \gamma_{k}\right)$, where $\gamma_{k}=\frac{2-\lambda_{k+1}}{2-\lambda_{k}}<1$, compute the approximate SE $\widetilde{Y}$ via the Power method:

1) $M \stackrel{\text { def }}{=} I+D^{-1 / 2} A D^{-1 / 2}$;
2) Let $\widetilde{U} \widetilde{\Sigma} \widetilde{V}^{\mathrm{T}}$ be the $S V D$ of $M^{p} S$; and
3) $\widetilde{Y} \stackrel{\text { def }}{=} \widetilde{U} \in \mathbb{R}^{n \times k}$.

Then, with probability at least $1-2 e^{-2 n}-3 \beta$, it holds that

$$
\left\|Y Y^{\mathrm{T}}-\tilde{Y} \tilde{Y}^{\mathrm{T}}\right\|_{F} \leqslant \varepsilon
$$

In Subsection 7.2.2, we establish technical lemmas that allows us to apply the proof technique developed in [BKG15, Theorem 6] for approximate $\mathrm{SE} \widetilde{Y}$, and to prove a similar statement for the approximate normalized SE $\widetilde{Y^{\prime}}$.

Theorem 5.5 (from page 62). Let $\varepsilon, \delta_{p} \in(0,1)$ be arbitrary. Compute the approximate normalized $S E$ $\widetilde{Y^{\prime}}$ via the Power method with $p \geqslant \ln \left(8 n k / \varepsilon \delta_{p}\right) / \ln \left(1 / \gamma_{k}\right)$ iterations and $\gamma_{k}=\left(2-\lambda_{k+1}\right) /\left(2-\lambda_{k}\right)<1$. Run on the rows of $\widetilde{Y^{\prime}}$ an $\alpha$-approximate $\bar{k}$-means clustering algorithm with failure probability $\delta_{\alpha}$. Let the outcome be a clustering indicator matrix $\widetilde{X_{\alpha}^{\prime}} \in \mathbb{R}^{n \times k}$. Then, with probability at least $1-2 e^{-2 n}-3 \delta_{p}-\delta_{\alpha}$, it holds that

$$
\left\|Y^{\prime}-\widetilde{X_{\alpha}^{\prime}}\left(\widetilde{X_{\alpha}^{\prime}}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F}^{2} \leqslant(1+4 \varepsilon) \cdot \alpha \cdot\left\|Y^{\prime}-X_{\mathrm{opt}}^{\prime}\left(X_{\mathrm{opt}}^{\prime}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F}^{2}+4 \varepsilon^{2}
$$

In Subsection 7.2.3, we prove Theorem 5.6 using Lemma 7.3 and Theorem 5.5, showing that the approximate normalized $\mathrm{SE} \widetilde{Y^{\prime}}$, computed via the Power method, is $\varepsilon$-separated.

Theorem 5.6 (from page 62). Assume $\Psi=20^{4} \cdot k^{3} / \delta, k / \delta \geqslant 10^{9}$ for some $\delta \in(0,1 / 2]$. Compute the approximate normalized $S E \widetilde{Y^{\prime}}$ via the Power method with $p \geqslant \Omega\left(\frac{\ln n}{\lambda_{k+1}}\right)$. Then, for $\varepsilon=6 \cdot 10^{-7}$ it holds with high probability that

$$
\triangle_{k}\left(\widetilde{Y^{\prime}}\right)<5 \varepsilon^{2} \cdot \triangle_{k-1}\left(\widetilde{Y^{\prime}}\right)
$$

In Subsection 7.3, we show that Part (b) of Theorem 5.2 follows by combining Part (a) of Theorem 5.2, Theorem 5.3, Theorem 5.5 and Theorem 5.6.

### 7.2.1 Proof of Lemma 7.3

We argue in a similar manner as in [BM14, Lemma 7]. Our analysis uses the following two probabilistic results on Gaussian matrices.

Lemma 7.4 (Norm of a Gaussian Matrix [DS01]). Let $M \in \mathbb{R}^{n \times k}$ be a matrix of i.i.d. standard Gaussian random variables, where $n \geqslant k$. Then, for $t \geqslant 4, \operatorname{Pr}\left\{\sigma_{1}(M) \geqslant t \sqrt{n}\right\} \geqslant \exp \left\{-n t^{2} / 8\right\}$.

Lemma 7.5 (Invertibility of a Gaussian Matrix [SST06]). Let $M \in \mathbb{R}^{n \times n}$ be a matrix of i.i.d. standard Gaussian random variables. Then, for any $\beta \in(0,1), \operatorname{Pr}\left\{\sigma_{n}(M) \leqslant \beta /(2.35 \sqrt{n})\right\} \leqslant \beta$.

Using the preceding two lemmas, we obtain the following probabilistic statement.
Lemma 7.6 (Rectangular Gaussian Matrix). Let $S \in \mathbb{R}^{n \times k}$ be a matrix of i.i.d. standard Gaussian random variables, $V \in \mathbb{R}^{n \times \rho}$ be a matrix with orthonormal columns and $n \geqslant \rho \geqslant k$. Then, with probability at least $1-e^{-2 n}$ it holds $\operatorname{rank}\left(V^{\mathrm{T}} S\right)=k$.

Proof. Let $S^{\prime} \in \mathbb{R}^{n \times \rho}$ be an extension of $S$ such that $S^{\prime}=\left[S S^{\prime \prime}\right]$, where $S^{\prime \prime} \in \mathbb{R}^{n \times \rho-k}$ is a matrix of i.i.d. standard Gaussian random variables. Notice that $V^{\mathrm{T}} S^{\prime} \in \mathbb{R}^{\rho \times \rho}$ is a matrix of i.i.d. standard Gaussian random variables. We apply now Lemma 7.5 with $\beta=e^{-2 n}$ which yields with probability at least $1-e^{-2 n}$ that $\sigma_{\rho}\left(V^{\mathrm{T}} S^{\prime}\right)>1 /\left(2.35 \cdot e^{2 n} \sqrt{\rho}\right)>0$ and thus $\operatorname{rank}\left(V^{\mathrm{T}} S^{\prime}\right)=\rho$. In particular, $\operatorname{rank}\left(V^{\mathrm{T}} S\right)=k$ with probability at least $1-e^{-2 n}$.

Proof of Lemma 7.3. By the Eigendecomposition theorem, $\mathcal{L}_{G}=U \Sigma^{\prime} U^{-1}$ where $U \in \mathbb{R}^{n \times n}$ is an orthonormal matrix whose $i$-th column equals the eigenvector of $\mathcal{L}_{G}$ corresponding to the $i$-th smallest eigenvalue $\lambda_{i}$, and $\Sigma^{\prime}$ is a non-negative diagonal matrix such that $\Sigma_{i i}^{\prime}=\lambda_{i}$, for all $i$. Since the canonical SE $Y \in \mathbb{R}^{n \times k}$ consists of the bottom $k$ eigenvectors of $\mathcal{L}_{G}$, we have $U=\left[Y U_{n-k}\right]$ where $U_{n-k} \in \mathbb{R}^{n \times n-k}$, and similarly $\Sigma=\left[\Sigma_{k} 0_{k, n-k} ; 0_{n-k, k} \Sigma_{n-k}\right]$.

Further, by the Eigendecomposition theorem $M=U \Sigma U^{\mathrm{T}}$, where $\Sigma=2 I-\Sigma^{\prime}$ and in particular $\Sigma_{i i}=2-\lambda_{i} \geqslant 0$ for all $i$. Since $M^{p}=U \Sigma^{p} U^{\mathrm{T}}$, it follows that $\operatorname{ker}\left(M^{p} S\right)=\operatorname{ker}\left(U^{\mathrm{T}} S\right)$. By Lemma 7.6 with probability at least $1-e^{-2 n}$ we have $\operatorname{rank}\left(U^{\mathrm{T}} S\right)=k$ and thus matrix $M^{p} S$ has $k$ singular values. Further, the SVD decomposition $\widetilde{U} \widetilde{\Sigma} \widetilde{V}^{\mathrm{T}}$ of $M^{p} S$ satisfies: $\widetilde{U} \in \mathbb{R}^{n \times k}$ is a matrix with orthonormal columns, $\widetilde{\Sigma} \in \mathbb{R}^{k \times k}$ is a positive diagonal matrix and $\widetilde{V}^{\mathrm{T}} \in \mathbb{R}^{k \times k}$ is an orthonormal matrix. Recall that approximate SE is defined by $\widetilde{Y}=\widetilde{U}$.

Let $R \stackrel{\text { def }}{=} \widetilde{\Sigma} \widetilde{V}^{\mathrm{T}} \in \mathbb{R}^{k \times k}$ and observe that $\widetilde{Y} R=M^{p} S=\left[\begin{array}{ll}Y & U_{n-k}\end{array}\right] \Sigma^{p}\left[Y^{\mathrm{T}} ; U_{n-k}^{\mathrm{T}}\right] S$. We use the facts:

$$
\begin{align*}
\tilde{Y} R & =Y \Sigma_{k}^{p} Y^{\mathrm{T}} S+U_{n-k} \Sigma_{n-k}^{p} U_{n-k}^{\mathrm{T}} S  \tag{7.5}\\
\sigma_{i}(\widetilde{Y} R) & \geqslant \sigma_{k}\left(Y \Sigma_{k}^{p} Y^{\mathrm{T}} S\right) \geqslant\left(2-\lambda_{k}\right)^{p} \cdot \sigma_{k}\left(Y^{\mathrm{T}} S\right) ;  \tag{7.6}\\
\sigma_{i}(\widetilde{Y} R) & =\sigma_{i}(R) ;  \tag{7.7}\\
\|X \widetilde{Y}\|_{2} & \geqslant\|X \widetilde{Y}\|_{2} \cdot \sigma_{k}(\widetilde{Y}), \quad \text { for any } X \in \mathbb{R}^{\ell \times k} . \tag{7.8}
\end{align*}
$$

(7.5) follows from the eigenvalue decomposition of $M$ and the fact that $M^{p}=U \Sigma^{p} U^{\mathrm{T}}$; (7.6) follows by (7.5) due to $Y$ and $U_{n-k}$ span orthogonal spaces, and since the minimum singular value of a product is at least the product of the minimum singular values; (7.7) holds due to $\widetilde{Y}^{\mathrm{T}} \widetilde{Y}=I_{k}$; Recall that with probability at least $1-e^{-2 n}$ we have $\sigma_{k}(R)>0$ and hence (7.8) follows by

$$
\|X\|_{2}=\max _{x \neq 0} \frac{\|X R x\|_{2}}{\|R x\|_{2}} \leqslant \max _{x \neq 0} \frac{\|X R x\|_{2}}{\sigma_{k}(R)\|x\|_{2}}=\frac{\|X R\|_{2}}{\sigma_{k}(R)} .
$$

[GVL12, Theorem 2.6.1] shows that for every two $m \times k$ orthonormal matrices $W, Z$ with $m \geqslant k$ it holds

$$
\left\|W W^{\mathrm{T}}-Z Z^{\mathrm{T}}\right\|_{2}=\left\|Z^{\mathrm{T}} W^{\perp}\right\|_{2}=\left\|W^{\mathrm{T}} Z^{\perp}\right\|_{2}
$$

where $\left[Z, Z^{\perp}\right] \in \mathbb{R}^{m \times m}$ is full orthonormal basis. Therefore, we have

$$
\begin{equation*}
\left\|Y Y^{\mathrm{T}}-\tilde{Y} \widetilde{Y}^{\mathrm{T}}\right\|_{2}=\left\|\widetilde{Y}^{\mathrm{T}} Y^{\perp}\right\|_{2}=\left\|\left(Y^{\perp}\right)^{\mathrm{T}} \tilde{Y}\right\|_{2}=\left\|U_{n-k}^{\mathrm{T}} \widetilde{Y}\right\|_{2} \tag{7.9}
\end{equation*}
$$

where the last equality is due to $Y^{\perp}=U_{n-k}$.
To upper bound $\left\|U_{n-k}^{\mathrm{T}} \widetilde{Y}\right\|_{2}$ we establish the following inequalities:

$$
\begin{align*}
& \left\|U_{n-k}^{\mathrm{T}} \tilde{Y} R\right\|_{2} \geqslant\left\|U_{n-k}^{\mathrm{T}} \widetilde{Y}\right\|_{2} \cdot \sigma(R) \geqslant\left\|U_{n-k}^{\mathrm{T}} \widetilde{Y}\right\|_{2} \cdot\left(2-\lambda_{k}\right)^{p} \cdot \sigma_{k}\left(Y^{\mathrm{T}} S\right)  \tag{7.10}\\
& \left\|U_{n-k}^{\mathrm{T}} \widetilde{Y} R\right\|_{2}=\left\|\Sigma_{n-k}^{p} U_{n-k}^{\mathrm{T}} S\right\|_{2} \leqslant\left(2-\lambda_{k+1}\right)^{p} \cdot \sigma_{1}\left(U_{n-k}^{\mathrm{T}} S\right) \tag{7.11}
\end{align*}
$$

where (7.10) follows by (7.8), (7.7) and (7.6); and (7.11) is due to (7.5) and $2=\Sigma_{11} \geqslant \cdots \geqslant \Sigma_{n n} \geqslant 0$.
By Lemma 7.4 and Lemma 7.5, it follows by the Union bound that with probability at least $1-e^{-2 n}-3 \beta$, we have

$$
\begin{equation*}
\frac{\beta}{\sqrt{k}} \leqslant \sigma_{k}\left(Y^{\mathrm{T}} S\right) \quad \text { and } \quad \sigma_{1}\left(U_{n-k}^{\mathrm{T}} S\right) \leqslant 4 \sqrt{n} \tag{7.12}
\end{equation*}
$$

Using (7.9), (7.10), (7.11) and (7.12) we obtain

$$
\begin{equation*}
\left\|Y Y^{\mathrm{T}}-\tilde{Y} \tilde{Y}^{\mathrm{T}}\right\|_{2}=\left\|U_{n-k}^{\mathrm{T}} \tilde{Y}\right\|_{2} \leqslant(4 / \beta) \cdot \sqrt{n k} \cdot \gamma_{k}^{p} \tag{7.13}
\end{equation*}
$$

Since $\|M\|_{F} \leqslant \sqrt{\operatorname{rank}(M)} \cdot\|M\|_{2}$ for every matrix $M$ and $\operatorname{rank}\left(Y Y^{\mathrm{T}}-\tilde{Y} \tilde{Y}^{\mathrm{T}}\right) \leqslant 2 k$, it follows

$$
\left\|Y Y^{\mathrm{T}}-\tilde{Y} \widetilde{Y}^{\mathrm{T}}\right\|_{F} \leqslant 2 k \cdot\left\|Y Y^{\mathrm{T}}-\tilde{Y} \widetilde{Y}^{\mathrm{T}}\right\|_{2} \leqslant(8 / \beta) \cdot n^{1 / 2} k^{3 / 2} \cdot \gamma_{k}^{p} \leqslant \varepsilon
$$

where the last two inequalities are due to (7.13) and the choice of $\gamma_{k}$.

### 7.2.2 Proof of Theorem 5.5

[BKG15, Theorem 6] relates canonical SE and approximate SE, whereas our goal is to establish similar result for the normalized SE and approximate normalized SE. We present next four technical lemmas that combined with Lemma 7.3, allow us to apply the proof technique developed in [BKG15, Theorem 6].

Lemma 7.7. Let $X^{\prime}, \widetilde{X^{\prime}} \in \mathbb{R}^{m \times k}$ be indicator matrices returned by an $\alpha$-approximate $k$-means clustering algorithm applied on inputs $Y^{\prime}$ and $\widetilde{Y}^{\prime}$, respectively, for any $\alpha \geqslant 1$. Then, it holds that $X^{\prime}\left(X^{\prime}\right)^{\mathrm{T}}$ and $\widetilde{X^{\prime}}\left(\widetilde{X^{\prime}}\right)^{\mathrm{T}}$ are projection matrices.

Proof. We prove now the first conclusion. By construction, there are $d(v)$ many copies of row $Y(v,:) / \sqrt{d(v)}$ in $Y^{\prime}$, for all $v \in V$. W.l.o.g. the indicator matrix $X^{\prime}$ has all copies of row $Y(v,:) / \sqrt{d(v)}$ assigned to the same cluster, for all $v \in V$. By definition, $X_{i j}^{\prime}=1 / \sqrt{\mu\left(C_{j}\right)}$ if row $Y_{i,:}^{\prime}$ belongs to the $j$-th cluster $C_{j}$ and $X_{i j}^{\prime}=0$ otherwise. Hence, $\left(X^{\prime}\right)^{\mathrm{T}} X^{\prime}=I_{k \times k}$ and thus $\left[X^{\prime}\left(X^{\prime}\right)^{\mathrm{T}}\right]^{2}=X^{\prime}\left(X^{\prime}\right)^{\mathrm{T}}$. The second part follows similarly, since matrix $\widetilde{U}$ is orthonormal.

Lemma 7.8. The normalized $S E Y^{\prime}$ and the approximate normalized $S E \widetilde{Y^{\prime}}$ are orthonormal matrices.
Proof. We prove now $\left(Y^{\prime}\right)^{\mathrm{T}} Y^{\prime}=I_{k \times k}$. The equality ${\widetilde{Y^{\prime}}}^{\mathrm{T}} \widetilde{Y^{\prime}}=I_{k \times k}$ follows similarly. Note that

$$
\begin{aligned}
{\left[\left(Y^{\prime}\right)^{\mathrm{T}} Y^{\prime}\right]_{i j} } & =\left(\begin{array}{lll}
\frac{Y(1, i)}{\sqrt{d(1)}} \mathbf{1}_{d(1)}^{\mathrm{T}} & \cdots & \frac{Y(n, i)}{\sqrt{d(n)}} \mathbf{1}_{d(n)}^{\mathrm{T}}
\end{array}\right)\left(\begin{array}{c}
\frac{Y(1, j)}{\sqrt{d(1)}} \mathbf{1}_{d(1)} \\
\cdots \\
\frac{Y(n, j)}{\sqrt{d(n)}} \mathbf{1}_{d(n)}
\end{array}\right) \\
& =\sum_{\ell=1}^{n} d(\ell) \frac{Y(\ell, i)}{\sqrt{d(\ell)}} \frac{Y(\ell, j)}{\sqrt{d(\ell)}}=\langle Y(:, i), Y(:, j)\rangle=\delta(i, j),
\end{aligned}
$$

where $\delta(i, j)$ is the Kronecker delta function. Hence, the statement follows.

Lemma 7.9. It holds that $\left\|Y^{\prime}\left(Y^{\prime}\right)^{\mathrm{T}}-\widetilde{Y^{\prime}}\left(\widetilde{Y^{\prime}}\right)^{\mathrm{T}}\right\|_{F}=\left\|Y Y^{\mathrm{T}}-\widetilde{Y} \widetilde{Y}^{\mathrm{T}}\right\|_{F}$.
Proof. Let $\mathbf{1}_{d(i)} \in\{0,1\}^{m}$ be an indicator vector of the $d(i)$ copies of row $Y(i,:) / \sqrt{d(i)}$ in matrix $Y^{\prime}$. By definition

$$
Y^{\prime}\left(Y^{\prime}\right)^{\mathrm{T}}=\sum_{\ell=1}^{k} Y_{:, \ell}^{\prime} Y_{:, \ell}^{\prime T} \quad \text { where } \quad Y_{:, \ell}^{\prime}=\left(\begin{array}{c}
\frac{Y(1, \ell)}{\sqrt{d(1)}} \mathbf{1}_{d(1)} \\
\cdots \\
\frac{Y(n, \ell)}{\sqrt{d(n)}} \mathbf{1}_{d(n)}
\end{array}\right)_{m \times 1}
$$

and

$$
\left(Y_{:, \ell}^{\prime} Y_{:, \ell}^{\prime T}\right)_{d(i) d(j)}=\frac{Y(i, \ell) Y(j, \ell)}{\sqrt{d(i) d(j)}} \cdot \mathbf{1}_{d(i)} \mathbf{1}_{d(j)}^{\mathrm{T}}
$$

Hence, we have

$$
\begin{aligned}
\left\|Y^{\prime}\left(Y^{\prime}\right)^{\mathrm{T}}-\widetilde{Y^{\prime}}\left(\widetilde{Y^{\prime}}\right)^{\mathrm{T}}\right\|_{F}^{2} & =\sum_{i=1}^{n} \sum_{j=1}^{n}\left\|\left(Y^{\prime}\left(Y^{\prime}\right)^{\mathrm{T}}-\widetilde{Y^{\prime}}\left(\widetilde{Y^{\prime}}\right)^{\mathrm{T}}\right)_{d(i) d(j)}\right\|_{F}^{2} \\
& =\sum_{i=1}^{n} \sum_{j=1}^{n}\left\|\sum_{\ell=1}^{k}\left(Y_{:, \ell}^{\prime}\left(Y_{:, \ell}^{\prime}\right)^{\mathrm{T}}-\widetilde{Y^{\prime}}:,\left(\widetilde{Y^{\prime}}:, \ell\right)^{\mathrm{T}}\right)_{d(i) d(j)}\right\|_{F}^{2} \\
& =\sum_{i=1}^{n} \sum_{j=1}^{n}\left\|\left\{\sum_{\ell=1}^{k}\left(\frac{Y(i, \ell) Y(j, \ell)}{\sqrt{d(i) d(j)}}-\frac{\widetilde{Y}(i, \ell) \widetilde{Y}(j, \ell)}{\sqrt{d(i) d(j)}}\right)\right\} \cdot \mathbf{1}_{d(i)} \mathbf{1}_{d(j)}^{\mathrm{T}}\right\|_{F}^{2} .
\end{aligned}
$$

By definition of Frobenius norm, it holds that

$$
\begin{aligned}
\left\|Y^{\prime}\left(Y^{\prime}\right)^{\mathrm{T}}-\widetilde{Y}^{\prime}\left(\widetilde{Y}^{\prime}\right)^{\mathrm{T}}\right\|_{F}^{2} & =\sum_{i=1}^{n} \sum_{j=1}^{n} d(i) d(j)\left[\sum_{\ell=1}^{k}\left(\frac{Y(i, \ell) Y(j, \ell)}{\sqrt{d(i) d(j)}}-\frac{\widetilde{Y}(i, \ell) \widetilde{Y}(j, \ell)}{\sqrt{d(i) d(j)}}\right)\right]^{2} \\
& =\sum_{i=1}^{n} \sum_{j=1}^{n}\left[\sum_{\ell=1}^{k}(Y(i, \ell) Y(j, \ell)-\widetilde{Y}(i, \ell) \widetilde{Y}(j, \ell))\right]^{2} \\
& =\sum_{i=1}^{n} \sum_{j=1}^{n}\left(Y Y^{\mathrm{T}}-\widetilde{Y} \widetilde{Y}^{\mathrm{T}}\right)_{i j}^{2} \\
& =\left\|Y Y^{\mathrm{T}}-\widetilde{Y} \widetilde{Y}^{\mathrm{T}}\right\|_{F}^{2} .
\end{aligned}
$$

Lemma 7.10. For any matrix $U$ with orthonormal columns and every matrix $A$ it holds

$$
\begin{equation*}
\left\|U U^{\mathrm{T}}-A A^{\mathrm{T}} U U^{\mathrm{T}}\right\|_{F}=\left\|U-A A^{\mathrm{T}} U\right\|_{F} \tag{7.14}
\end{equation*}
$$

Proof. The statement follows by the Frobenius norm property $\|M\|_{F}^{2}=\operatorname{Tr}\left[M^{\mathrm{T}} M\right]$, the cyclic property of trace $\operatorname{Tr}\left[U M^{\mathrm{T}} M U^{\mathrm{T}}\right]=\operatorname{Tr}\left[M^{\mathrm{T}} M \cdot U^{\mathrm{T}} U\right]$ and the orthogonality of matrix $U$.

Using the preceding lemmas, we are ready to prove Theorem 5.5.
Proof of Theorem 5.5. Using Lemma 7.3 and Lemma 7.9 with probability at least $1-2 e^{-2 n}-3 \delta_{p}$ we have

$$
\left\|Y^{\prime}\left(Y^{\prime}\right)^{\mathrm{T}}-\widetilde{Y^{\prime}}\left(\widetilde{Y^{\prime}}\right)^{\mathrm{T}}\right\|_{F}=\left\|Y Y^{\mathrm{T}}-\widetilde{Y} \widetilde{Y}^{\mathrm{T}}\right\|_{F} \leqslant \varepsilon .
$$

Let $Y^{\prime}\left(Y^{\prime}\right)^{\mathrm{T}}=\widetilde{Y^{\prime}}\left(\widetilde{Y^{\prime}}\right)^{\mathrm{T}}+E$ such that $\|E\|_{F} \leqslant \varepsilon$. By combining Lemma 7.8 and Lemma 7.10, (7.14) holds for the matrices $Y^{\prime}$ and $\widetilde{Y^{\prime}}$. Thus, by Lemma 7.7 and the proof techniques in [BKG15, Theorem 6], it follows that

$$
\begin{equation*}
\left\|Y^{\prime}-\widetilde{X_{\alpha}^{\prime}}\left(\widetilde{X_{\alpha}^{\prime}}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F} \leqslant \sqrt{\alpha} \cdot\left(\left\|Y^{\prime}-X_{\mathrm{opt}}^{\prime}\left(X_{\mathrm{opt}}^{\prime}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F}+2 \varepsilon\right) . \tag{7.15}
\end{equation*}
$$

The desired statement follows by simple algebraic manipulations of (7.15).

### 7.2.3 Proof of Theorem 5.6

In this section, we demonstrate that the approximate normalized $\mathrm{SE} \widetilde{Y^{\prime}}$ is $\varepsilon$-separated, i.e. $\triangle_{k}\left(\widetilde{Y^{\prime}}\right)<$ $5 \varepsilon^{2} \cdot \triangle_{k-1}\left(\widetilde{Y^{\prime}}\right)$. Our analysis builds upon Theorem 5.4, Theorem 5.5 and the proof techniques in [BKG15, Theorem 6].

Before we present the proof of Theorem 5.6, we establish two technical Lemmas.
Lemma 7.11. Suppose $\Psi \geqslant 20^{4} \cdot k^{3} / \delta$ for some $\delta \in(0,1 / 2]$. Then, it holds that

$$
\ln \left(\frac{2-\lambda_{k}}{2-\lambda_{k+1}}\right) \geqslant \frac{1}{2}\left(1-\frac{4 \delta}{20^{4} k^{2}}\right) \lambda_{k+1}
$$

Proof. By (5.5), the following higher-order Cheeger inequalities hold

$$
\begin{equation*}
\lambda_{k} / 2 \leqslant \rho(k) \leqslant O\left(k^{2}\right) \cdot \sqrt{\lambda_{k}} \tag{7.16}
\end{equation*}
$$

Using the LHS of (7.16), we have

$$
k^{3} \widehat{\rho}_{\mathrm{avr}}(k)=k^{2} \sum_{i=1}^{k} \phi\left(P_{i}\right) \geqslant k^{2} \max _{i \in\{1, \ldots, k\}} \phi\left(P_{i}\right) \geqslant k^{2} \cdot \rho(k) \geqslant \frac{k^{2} \lambda_{k}}{2}
$$

and thus the $k$-th smallest eigenvalue of $\mathcal{L}_{G}$ satisfies $\lambda_{k} \leqslant 2 k \cdot \widehat{\rho}_{\text {avr }}(k)$. Then, the gap assumption yields

$$
\lambda_{k+1} \geqslant \frac{20^{4} k^{2}}{2 \delta} \cdot 2 k \cdot \widehat{\rho}_{\mathrm{avr}}(k) \geqslant \frac{20^{4} k^{2}}{2 \delta} \cdot \lambda_{k}
$$

The statement follows by

$$
\frac{2-\lambda_{k}}{2-\lambda_{k+1}} \geqslant \frac{1-\frac{\delta}{20^{4} k^{2}} \cdot \lambda_{k+1}}{1-\frac{1}{2} \lambda_{k+1}} \geqslant \exp \left\{\frac{1}{2}\left(1-\frac{4 \delta}{20^{4} k^{2}}\right) \lambda_{k+1}\right\}
$$

Lemma 7.12. For any matrices $A \in \mathbb{R}^{m \times n}$ and $B \in \mathbb{R}^{n \times k}$, it holds that $\|A B\|_{F} \leqslant\|A\|_{2} \cdot\|B\|_{F}$.
Proof. By definition, $\|B\|_{F}^{2}=\sum_{i=1}^{k}\left\|B_{:, i}\right\|_{2}^{2}$ and $\|A x\|_{2} \leqslant\|A\|_{2}\|x\|_{2}$, and thus we have

$$
\|A B\|_{F}^{2}=\sum_{i=1}^{k}\left\|A B_{:, i}\right\|_{2}^{2} \leqslant\|A\|_{2}^{2} \sum_{i=1}^{k}\left\|B_{:, i}\right\|_{2}^{2}=\|A\|_{2}^{2} \cdot\|B\|_{2}^{2}
$$

In the following, we use interchangeably $X_{\mathrm{opt}}^{\prime}$ and $X_{\mathrm{opt}}^{\prime(k)}$ to denote an optimal indicator matrix for the $k$-means clustering problem on $Y^{\prime}$. Similarly, we denote by $X_{\mathrm{opt}}^{\prime(k-1)}$ an optimal indicator matrix for the $(k-1)$-means clustering problem on $Y^{\prime}$.

We are now ready to prove Theorem 5.6.

Proof of Theorem 5.6. We set the approximation parameter in Theorem 5.5 to $\varepsilon^{\prime}=\varepsilon / 30$. By Theorem 5.4, we have

$$
\begin{align*}
& \left\|Y^{\prime}-X_{\mathrm{opt}}^{\prime(k)}\left(X_{\mathrm{opt}}^{\prime(k)}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F}=\sqrt{\triangle_{k}\left(Y^{\prime}\right)} \\
\leqslant & \varepsilon \sqrt{\triangle_{k-1}\left(Y^{\prime}\right)}=\varepsilon\left\|Y^{\prime}-X_{\mathrm{opt}}^{\prime(k-1)}\left(X_{\mathrm{opt}}^{\prime(k-1)}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F} \tag{7.17}
\end{align*}
$$

We compute an approximate $\operatorname{SE} \widetilde{Y} \in \mathbb{R}^{n \times k}$, defined in (5.7), via the Power method which runs $p=O\left(\frac{\ln n}{\lambda_{k+1}}\right)$ iterations.

By combining Lemma 7.3 and Lemma 7.9, for the normalized and approximate normalized $\mathrm{SE}, Y^{\prime}$ and $\widetilde{Y}^{\prime}$ respectively, we obtain w.h.p. that

$$
\left\|Y^{\prime} Y^{\prime \mathrm{T}}-\widetilde{Y^{\prime}}\left(\widetilde{Y^{\prime}}\right)^{\mathrm{T}}\right\|_{F}=\left\|Y Y^{\mathrm{T}}-\widetilde{Y} \widetilde{Y}^{\mathrm{T}}\right\|_{F} \leqslant \varepsilon^{\prime}
$$

Let $Y^{\prime} Y^{\prime \mathrm{T}}=\widetilde{Y^{\prime}}{\widetilde{Y^{\prime}}}^{\mathrm{T}}+E$ such that $\|E\|_{F} \leqslant \varepsilon^{\prime}$. By Lemma $7.8, Y^{\prime}$ and $\widetilde{Y^{\prime}}$ are orthonormal matrices. Hence, by Lemma 7.10 applied on $\widetilde{Y^{\prime}}$, we obtain

$$
\begin{align*}
\sqrt{\triangle_{k}\left(\widetilde{Y^{\prime}}\right)} & =\left\|\widetilde{Y^{\prime}}-\widetilde{X_{\mathrm{opt}}^{\prime(k)}}\left(\widetilde{\left(X_{\mathrm{opt}}^{\prime(k)}\right.}\right)^{\mathrm{T}} \widetilde{Y^{\prime}}\right\|_{F}=\left\|\widetilde{Y^{\prime}}{\widetilde{Y^{\prime}}}^{\mathrm{T}}-\widetilde{X_{\mathrm{opt}}^{\prime(k)}}\left(\widetilde{\left(X_{\mathrm{opt}}^{\prime(k)}\right.}\right)^{\mathrm{T}}{\widetilde{Y^{\prime}}{\widetilde{Y^{\prime}}}^{\mathrm{T}} \|_{F}}=\right\| Y^{\prime}\left(Y^{\prime}\right)^{\mathrm{T}}-\widetilde{X_{\mathrm{opt}}^{\prime(k)}}\left(\widetilde{X_{\mathrm{opt}}^{\prime(k)}}\right)^{\mathrm{T}} Y^{\prime}\left(Y^{\prime}\right)^{\mathrm{T}}-\left[I_{m \times m}-\widetilde{X_{\mathrm{opt}}^{\prime(k)}}\left(\widetilde{X_{\mathrm{opt}}^{\prime(k)}}\right)^{\mathrm{T}}\right] E \|_{F} \\
& \leqslant\left\|Y^{\prime}-\widetilde{X_{\mathrm{opt}}^{\prime(k)}}\left(\widetilde{\left(X_{\mathrm{opt}}^{\prime(k)}\right.}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F}+\|E\|_{F},
\end{align*}
$$

where the last step uses triangle inequality, Lemma 7.10 applied on $Y^{\prime}$, Lemma 7.12, Lemma 7.7 and $\left\|I-P P^{\mathrm{T}}\right\|_{2} \leqslant 1$ for any projection matrix $P$. Then, we apply Theorem 5.5 with an exact $k$-means clustering algorithm, i.e. $\alpha=1$, and parameters $\delta_{p}=n^{-O(1)}$ and $\varepsilon^{\prime}$ (as above). By Lemma 7.11, for any $p \geqslant \Omega\left(\frac{\ln n}{\lambda_{k+1}}\right)$ it holds w.h.p.

$$
\begin{equation*}
\left\|Y^{\prime}-\widetilde{X_{\mathrm{opt}}^{\prime(k)}}\left(\widetilde{X_{\mathrm{opt}}^{\prime(k)}}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F}^{2} \leqslant\left(1+4 \varepsilon^{\prime}\right) \cdot\left\|Y^{\prime}-X_{\mathrm{opt}}^{\prime(k)}\left(X_{\mathrm{opt}}^{\prime(k)}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F}^{2}+4\left(\varepsilon^{\prime}\right)^{2} \tag{7.19}
\end{equation*}
$$

The proof proceeds by case distinction.
Case 1: Suppose $\varepsilon^{\prime} \leqslant \frac{1}{4} \sqrt{\triangle_{k}\left(Y^{\prime}\right)}$. Combining (7.18), (7.19) and $\|E\|_{F} \leqslant \varepsilon^{\prime}$, yields

$$
\begin{align*}
\sqrt{\triangle_{k}\left(\widetilde{Y^{\prime}}\right)} & \leqslant \varepsilon^{\prime}+\sqrt{\left(1+4 \varepsilon^{\prime}\right) \cdot \triangle_{k}\left(Y^{\prime}\right)+4\left(\varepsilon^{\prime}\right)^{2}} \\
& \leqslant 2 \cdot \sqrt{\triangle_{k}\left(Y^{\prime}\right)} \leqslant 2 \varepsilon \cdot \sqrt{\triangle_{k-1}\left(Y^{\prime}\right)} \tag{7.20}
\end{align*}
$$

where the last inequality follows by (7.17). Moreover, it holds that

$$
\begin{align*}
& \sqrt{\triangle_{k-1}\left(Y^{\prime}\right)}=\left\|Y^{\prime}-X_{\mathrm{opt}}^{\prime(k-1)}\left(X_{\mathrm{opt}}^{\prime(k-1)}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F} \leqslant\left\|Y^{\prime}-\widetilde{X_{\mathrm{opt}}^{\prime(k-1)}}\left(\widetilde{\left(X_{\mathrm{opt}}^{\prime(k-1)}\right.}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F} \\
& =\left\|Y^{\prime}\left(Y^{\prime}\right)^{\mathrm{T}}-\widetilde{X_{\mathrm{opt}}^{\prime(k-1)}}\left(\widetilde{X_{\mathrm{opt}}^{\prime(k-1)}}\right)^{\mathrm{T}} Y^{\prime}\left(Y^{\prime}\right)^{\mathrm{T}}\right\|_{F} \\
& =\left\|\widetilde{Y^{\prime} \widetilde{Y}^{\prime}}{ }^{\mathrm{T}}-\widetilde{X_{\mathrm{opt}}^{\prime(k-1)}}\left(\widetilde{X_{\mathrm{opt}}^{\prime(k-1)}}\right)^{\mathrm{T}} \widetilde{Y^{\prime} \widetilde{Y}^{\prime}}{ }^{\mathrm{T}}+\left[I_{m \times m}-\widetilde{X_{\mathrm{opt}}^{\prime(k-1)}}\left(\widetilde{X_{\mathrm{opt}}^{\prime(k-1)}}\right)^{\mathrm{T}}\right] E\right\|_{F} \\
& \leqslant\left\|\widetilde{Y^{\prime}}-\widetilde{X_{\mathrm{opt}}^{\prime(k-1)}}\left(\widetilde{X_{\mathrm{opt}}^{\prime(k-1)}}\right)^{\mathrm{T}} \widetilde{Y^{\prime}}\right\|_{F}+\|E\|_{F}  \tag{7.21}\\
& \leqslant \sqrt{\triangle_{k-1}\left(\widetilde{Y^{\prime}}\right)}+\frac{\varepsilon}{4} \sqrt{\triangle_{k-1}\left(Y^{\prime}\right)},
\end{align*}
$$

where the last inequality uses the definition of $\triangle_{k-1}\left(\widetilde{Y^{\prime}}\right)$ and

$$
\|E\|_{F} \leqslant \varepsilon^{\prime} \leqslant \frac{1}{4} \sqrt{\triangle_{k}\left(Y^{\prime}\right)} \leqslant \frac{\varepsilon}{4} \sqrt{\triangle_{k-1}\left(Y^{\prime}\right)}
$$

Hence,

$$
\begin{equation*}
\sqrt{\triangle_{k-1}\left(Y^{\prime}\right)} \leqslant\left(1+\frac{\varepsilon}{2}\right) \sqrt{\triangle_{k-1}\left(\widetilde{Y^{\prime}}\right)} . \tag{7.22}
\end{equation*}
$$

Therefore, by combining (7.20) and (7.22), we obtain

$$
\sqrt{\triangle_{k}\left(\widetilde{Y^{\prime}}\right)} \leqslant 2 \varepsilon \cdot \sqrt{\triangle_{k-1}\left(Y^{\prime}\right)} \leqslant(2+\varepsilon) \cdot \varepsilon \cdot \sqrt{\triangle_{k-1}\left(\widetilde{Y^{\prime}}\right)} .
$$

Case 2: Suppose $0 \leqslant \frac{1}{4} \sqrt{\triangle_{k}\left(Y^{\prime}\right)}<\varepsilon^{\prime}$. Combining (7.18), (7.19) and $\|E\|_{F} \leqslant \varepsilon^{\prime}$, yields

$$
\begin{equation*}
\sqrt{\triangle_{k}\left(\widetilde{Y^{\prime}}\right)} \leqslant \varepsilon^{\prime}+\sqrt{\left(1+4 \varepsilon^{\prime}\right) \cdot \triangle_{k}\left(Y^{\prime}\right)+4\left(\varepsilon^{\prime}\right)^{2}} \leqslant 6 \varepsilon^{\prime} \tag{7.23}
\end{equation*}
$$

Further, by Lemma 7.1 and $k / \delta \geqslant 10^{9}$ we have $\triangle_{k-1}\left(Y^{\prime}\right) \geqslant 1 / 13$, and thus using (7.21) we obtain

$$
\begin{equation*}
\sqrt{\triangle_{k-1}\left(\widetilde{Y^{\prime}}\right)} \geqslant \sqrt{\triangle_{k-1}\left(Y^{\prime}\right)}-\|E\|_{F} \geqslant 1 / 4-\varepsilon^{\prime} \geqslant 1 / 5 . \tag{7.24}
\end{equation*}
$$

By combining Equations (7.23) and (7.24), it follows that

$$
\sqrt{\triangle_{k}\left(\widetilde{Y^{\prime}}\right)} \leqslant 6 \varepsilon^{\prime} \leqslant 30 \varepsilon^{\prime} \cdot \sqrt{\triangle_{k-1}\left(\widetilde{Y^{\prime}}\right)}=\varepsilon \cdot \sqrt{\triangle_{k-1}\left(\widetilde{Y^{\prime}}\right)}
$$

### 7.3 Proof of Approximate Spectral Clustering

We prove now Part (b) of Theorem 5.2. Let $p=\Theta\left(\frac{\ln n}{\lambda_{k+1}}\right)$. We compute the matrix $M^{p} S$ in time $O(m k p)$ and its singular value decomposition $\widetilde{U} \widetilde{\Sigma} \widetilde{V}^{\mathrm{T}}$ in time $O\left(n k^{2}\right)$. Based on it, we construct in time $O(m k)$ the approximate normalized $\operatorname{SE} \widetilde{Y^{\prime}}$, see (5.8).

By Theorem 5.6, $\widetilde{Y^{\prime}}$ is $\varepsilon$-separated for $\varepsilon=6 \cdot 10^{-7}$, i.e. $\triangle_{k}\left(\widetilde{Y^{\prime}}\right)<5 \varepsilon^{2} \cdot \triangle_{k-1}\left(\widetilde{Y^{\prime}}\right)$. Let $\alpha=1+10^{-10}$. Then, by Theorem 5.3, there is an algorithm that outputs in time $O\left(m k^{2}+k^{4}\right)$ a $k$-way vector partition with indicator matrix $\widetilde{X_{\alpha}^{\prime}}$ such that with probability at least $1-O(\sqrt{\varepsilon})$, we have

$$
\left\|\widetilde{Y^{\prime}}-\widetilde{X_{\alpha}^{\prime}}\left(\widetilde{X_{\alpha}^{\prime}}\right)^{\mathrm{T}} \widetilde{Y^{\prime}}\right\|_{F}^{2} \leqslant\left(1+\frac{1}{10^{10}}\right) \cdot\left\|\widetilde{Y^{\prime}}-\widetilde{X_{\mathrm{opt}}^{\prime}}\left(\widetilde{X_{\mathrm{opt}}^{\prime}}\right)^{\mathrm{T}} \widetilde{Y^{\prime}}\right\|_{F}^{2} .
$$

Let $\eta \in\left(n^{-O(1)}, 1\right)$ be a parameter to be determined soon. Observe that for $\Psi=20^{4} \cdot k^{3} / \delta, \delta \in(0,1 / 2]$ and $k / \delta \geqslant 10^{9}$, by Equations (5.11), (6.3) and Lemma 6.6, it holds

$$
\begin{equation*}
\left\|Y^{\prime}-X_{\mathrm{opt}}^{\prime(k)}\left(X_{\mathrm{opt}}^{\prime(k)}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F}^{2} \leqslant \operatorname{Cost}\left(\left\{P_{i}, p^{(i)}\right\}_{i=1}^{k}\right) \leqslant \frac{2 k^{2}}{\Psi} \leqslant \frac{1}{8 \cdot 10^{13}} . \tag{7.25}
\end{equation*}
$$

Combining Theorem 5.6 and Equation (7.25), yields

$$
\frac{1}{n^{O(1)}} \leqslant\left\|Y^{\prime}-X_{\mathrm{opt}}^{\prime}\left(X_{\mathrm{opt}}^{\prime}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F} \leqslant \frac{1}{10^{6}}
$$

Using Lemma 7.11, we apply Theorem 5.5 with $\delta_{p}=n^{-O(1)}, \alpha=1+10^{-10}, \delta_{\alpha}=O(\sqrt{\varepsilon})$ and

$$
\varepsilon^{\prime}=\frac{\sqrt{\eta}}{4} \cdot \frac{1}{n^{O(1)}} \leqslant \frac{\sqrt{\eta}}{4} \cdot\left\|Y^{\prime}-X_{\mathrm{opt}}^{\prime}\left(X_{\mathrm{opt}}^{\prime}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F}
$$

and obtain with constant probability (close to 1 ) that

$$
\begin{aligned}
& \left\|\widetilde{Y^{\prime}}-\widetilde{X_{\alpha}^{\prime}}\left(\widetilde{X_{\alpha}^{\prime}}\right)^{\mathrm{T}} \widetilde{Y^{\prime}}\right\|_{F}^{2} \leqslant\left(1+4 \varepsilon^{\prime}\right) \cdot \alpha \cdot\left\|Y^{\prime}-X_{\mathrm{opt}}^{\prime}\left(X_{\mathrm{opt}}^{\prime}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F}^{2}+4 \varepsilon^{\prime 2} \\
= & {\left[\left(1+\sqrt{\eta}\left\|Y^{\prime}-X_{\mathrm{opt}}^{\prime}\left(X_{\mathrm{opt}}^{\prime}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F}\right) \alpha+\frac{\eta}{4}\right] \cdot\left\|Y^{\prime}-X_{\mathrm{opt}}^{\prime}\left(X_{\mathrm{opt}}^{\prime}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F}^{2} } \\
\leqslant & {\left[\left(1+\frac{\sqrt{\eta}}{10^{6}}\right) \cdot\left(1+\frac{1}{10^{10}}\right)+\frac{\eta}{4}\right] \cdot\left\|Y^{\prime}-X_{\mathrm{opt}}^{\prime}\left(X_{\mathrm{opt}}^{\prime}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F}^{2}, }
\end{aligned}
$$

Then, for $\eta=1 / 10^{6}$ the approximate solution $\widetilde{X_{\alpha}^{\prime}}$ yields a multiplicative approximation, satisfying

$$
\left\|Y^{\prime}-\widetilde{X_{\alpha}^{\prime}}\left(\widetilde{X_{\alpha}^{\prime}}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F}^{2} \leqslant\left(1+\frac{1}{10^{6}}\right)\left\|Y^{\prime}-X_{\mathrm{opt}}^{\prime}\left(X_{\mathrm{opt}}^{\prime}\right)^{\mathrm{T}} Y^{\prime}\right\|_{F}^{2}
$$

The statement follows by Part (a) of Theorem 5.2 applied to the $k$-way partition $\left(A_{1}, \ldots, A_{k}\right)$ of $V$ that is induced by the indicator matrix $\widetilde{X_{\alpha}^{\prime}}$.

## Appendix

Let $G$ be an undirected and unweighted graph consisting of $k$ cliques each of size $n / k$, and let $A$ be its adjacency matrix. Let $G^{R}$ be the graph $G$, plus a set $E_{R}$ of $k$ additional edges connecting the $k$ cliques in the form of a ring such that no two edges in $E_{R}$ (of the ring) share a vertex. In other words, let $A_{R}$ be the adjacency matrix of $E_{R}$, then $A_{R} A_{R}=S$ where $S$ is a diagonal matrix defined as $S_{i i}=1$ if vertex $i$ is connected to a vertex in a neighboring clique, and $S_{i i}=0$ otherwise.

Recall that the normalized Laplacian matrix of $G$ is given by $\mathcal{L}_{G}=I-\mathcal{A}$, where $\mathcal{A}=D^{-1 / 2} A D^{-1 / 2}$ and $D$ is a positive diagonal matrix with $D_{i i}=\operatorname{deg}(i)$, for all $i$. Then, the normalized Laplacian matrix of $G^{R}$ is defined as

$$
\mathcal{L}_{G^{R}}=I-(D+S)^{-1 / 2}\left(A+A_{R}\right)(D+S)^{-1 / 2}
$$

Our analysis relies on the following four lemmas. We note that the first lemma is folklore.
Lemma 7.13. It holds that $\lambda_{1}\left(\mathcal{L}_{G}\right)=\cdots=\lambda_{k}\left(\mathcal{L}_{G}\right)=0$ and $\lambda_{k+1}\left(\mathcal{L}_{G}\right)=\cdots=\lambda_{n}\left(\mathcal{L}_{G}\right)=1$. Moreover, $\lambda_{1}(\mathcal{A})=\cdots=\lambda_{k}(\mathcal{A})=1$ and $\lambda_{k+1}(\mathcal{A})=\cdots=\lambda_{k+1}(\mathcal{A})=0$. In particular, $\mathcal{A}$ is a symmetric positive semi-definite matrix.

Lemma 7.14. For any symmetric matrix $M \in \mathbb{R}^{n \times n}$ and any positive diagonal matrix $S \in \mathbb{R}^{n \times n}$, it holds that $\lambda_{i}\left(M S^{-1}\right)=\lambda_{i}\left(S^{-1 / 2} M S^{-1 / 2}\right)$ for every $i \in\{1, \ldots, n\}$.

Proof. Let $v_{i} \in \mathbb{R}^{n}$ and $\lambda_{i} \in \mathbb{R}$ be such that $M S^{-1} v_{i}=\lambda_{i} v_{i}$. Let $v_{i}=S^{1 / 2} y_{i}$ for some vector $y_{i} \in \mathbb{R}^{n}$ (it exists as $S$ is a positive diagonal matrix), then we have $M S^{-1} \cdot S^{1 / 2} y_{i}=\lambda_{i} \cdot S^{1 / 2} y_{i}$ and by multiplying from the left with $S^{-1 / 2}$, we obtain $S^{-1 / 2} M S^{-1 / 2} \cdot y_{i}=\lambda_{i} \cdot y_{i}$.

Lemma 7.15. [Kly00, Eig15] Let $Q \in \mathbb{R}^{n \times n}$ be a symmetric positive semi-definite matrix and $S \in \mathbb{R}^{n \times n}$ be a positive diagonal matrix. Let $\lambda_{1} \geqslant \ldots \geqslant \lambda_{n}$ and $\mu_{1} \geqslant \ldots \geqslant \mu_{n}$ be the eigenvalues of $Q$ and $Q S^{-1}$, respectively. Then, it holds that $\lambda_{i} \cdot \min _{j}\left\{S_{j j}^{-1}\right\} \leqslant \mu_{i} \leqslant \lambda_{i} \cdot \max _{j}\left\{S_{j j}^{-1}\right\}$.

Lemma 7.16. [Ste90, Corollary 4.10] Let $A, E \in \mathbb{R}^{n \times n}$ be arbitrary symmetric matrices. Then, their eigenvalues satisfy $\left|\lambda_{i}(A+E)-\lambda_{i}(A)\right| \leqslant\|E\|_{2}$ for every $i \in\{1, \ldots, n\}$.

Using the preceding four lemmas, we establish a lower bound on the eigenvalue $\lambda_{k+1}\left(\mathcal{L}_{G^{R}}\right)$.

Theorem 7.17. (Ring of $k$ Cliques) The $(k+1)$-st smallest eigenvalue of the normalized Laplacian matrix of graph $G^{R}$ satisfies $\lambda_{k+1}\left(\mathcal{L}_{G^{R}}\right) \geqslant 1-k / n$.

Proof. Let $Z \stackrel{\text { def }}{=} D^{-1} \cdot(D+S)$. Note that $Z$ is a positive diagonal matrix. By definition, we have

$$
\begin{aligned}
\mathcal{L}_{G^{R}} & =I-(D+S)^{-1 / 2}\left(A+A_{R}\right)(D+S)^{-1 / 2} \\
& =I-Z^{-1 / 2} \cdot D^{-1 / 2}\left(A+A_{R}\right) D^{-1 / 2} \cdot Z^{-1 / 2} .
\end{aligned}
$$

It suffices to upper bound the $(k+1)$-st largest eigenvalue of matrix $Z^{-1 / 2} \cdot D^{-1 / 2}\left(A+A_{R}\right) D^{-1 / 2} \cdot Z^{-1 / 2}$. Recall that $\mathcal{A}=D^{-1 / 2} A D^{-1 / 2}$ and thus

$$
Z^{-1 / 2} \cdot D^{-1 / 2}\left(A+A_{R}\right) D^{-1 / 2} \cdot Z^{-1 / 2}=Z^{-1 / 2} \mathcal{A} Z^{-1 / 2}+Z^{-1 / 2} \cdot D^{-1 / 2} A_{R} D^{-1 / 2} \cdot Z^{-1 / 2}
$$

By definition, $\|M\|_{2}^{2}=\max _{\|x\|_{2}=1} x^{T} M^{T} M x$ and since $A_{R} A_{R}=S$, we have

$$
\begin{equation*}
\left\|A_{R}\right\|_{2}^{2}=\max _{\|x\|_{2}=1} x^{T} S x=\max _{\|x\|_{2}=1} \sum_{j=1}^{2 k} x_{i_{j}}^{2} \leqslant 1 \tag{7.26}
\end{equation*}
$$

and

$$
\begin{equation*}
\left\|(D+S)^{-1}\right\|_{2}^{2}=\max _{\|x\|_{2}=1} x^{T}(D+S)^{-2} x \leqslant \max _{i}\left\{\frac{1}{\left(D_{i i}+S_{i i}\right)^{2}}\right\} \cdot \max _{\|x\|_{2}=1} x^{T} x \leqslant\left(\frac{k}{n}\right)^{2} \tag{7.27}
\end{equation*}
$$

In order to apply Lemma 7.16, we upper bound first the expression

$$
\begin{align*}
& \left\|Z^{-1 / 2} \cdot D^{-1 / 2} A_{R} D^{-1 / 2} \cdot Z^{-1 / 2}\right\|_{2} \stackrel{\text { Lem. }}{=}{ }^{7.14}\left\|A_{R} \cdot D^{-1} Z^{-1}\right\|_{2} \\
= & \left\|A_{R} \cdot(D+S)^{-1}\right\|_{2} \leqslant\left\|A_{R}\right\|_{2} \cdot\left\|(D+S)^{-1}\right\|_{2} \stackrel{(7.26) \text { and }}{\leqslant}(7.27)  \tag{7.28}\\
\leqslant & \frac{k}{n} .
\end{align*}
$$

Since $\mathcal{A}$ is a symmetric positive semi-definite matrix and $Z$ is a positive diagonal matrix, we have

$$
\begin{array}{r}
\lambda_{k+1}\left(Z^{-1 / 2} \mathcal{A} Z^{-1 / 2}\right) \stackrel{\text { Lem. } 7.14}{=} \lambda_{k+1}\left(\mathcal{A} Z^{-1}\right) \\
\stackrel{\text { Lem. }}{\leqslant} \lambda_{k+1}(\mathcal{A}) \cdot \max _{i}\left\{Z_{i i}^{-1}\right\}=\lambda_{k+1}(\mathcal{A}) \stackrel{\text { Lem. }}{=}{ }^{7.13} 0 . \tag{7.29}
\end{array}
$$

Therefore, the largest $(k+1)$-st eigenvalue

$$
\begin{aligned}
& \quad \lambda_{k+1}\left(Z^{-1 / 2} \cdot D^{-1 / 2}\left(A+A_{R}\right) D^{-1 / 2} \cdot Z^{-1 / 2}\right) \\
& \stackrel{\text { Lem. }}{\leqslant}{ }^{7.16} \lambda_{k+1}\left(Z^{-1 / 2} \mathcal{A} Z^{-1 / 2}\right)+\left\|Z^{-1 / 2} \cdot D^{-1 / 2} A_{R} D^{-1 / 2} \cdot Z^{-1 / 2}\right\|_{2} \\
& \stackrel{\text { (7.28) and }}{\leqslant}(7.29) \\
&
\end{aligned}
$$

and thus the smallest $(k+1)$-st eigenvalue of $\mathcal{L}_{G^{R}}$ satisfies $\lambda_{k+1}\left(\mathcal{L}_{G^{R}}\right) \geqslant 1-k / n$.

## Part III

## Two Results on Slime Mold Computations

## Proceedings

This part is based on an article $\left[\mathrm{BBK}^{+} 18\right]$ that appeared at the journal of Theoretical Computer Science, Series C (Theory of Natural Computing). The full ArXiv version is available online $\left[\mathrm{BBK}^{+} 17\right]$.

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## Chapter 8

## Introduction

We present two results on slime mold computations, one on the biologically-grounded model and one on the biologically-inspired model. The former model was introduced by biologists to capture the slime's apparent ability to compute shortest paths. We show in Section 8.1 that the dynamics can actually do more. It can solve a wide class of linear programs with nonnegative cost vectors. The latter model was designed as an optimization technique inspired by the former model. We present in Section 8.2 an improved convergence result for its discretization.

### 8.1 The Biologically-Grounded Model

Physarum polycephalum is a slime mold that apparently is able to solve shortest path problems. Nakagaki, Yamada, and Tóth [NYT00] report about the following experiment; see Figure 8.1. They built a maze, covered it by pieces of Physarum (the slime can be cut into pieces which will reunite if brought into vicinity), and then fed the slime with oatmeal at two locations. After a few hours the slime retracted to a path that follows the shortest path in the maze connecting the food sources. The authors report that they repeated the experiment with different mazes; in all experiments, Physarum retracted to the shortest path.

The paper [TKN07] proposes a mathematical model for the behavior of the slime and argues extensively that the model is adequate. Physarum is modeled as an electrical network with time varying resistors. We have a simple undirected graph $G=(N, E)$ with distinguished nodes $s_{0}$ and $s_{1}$ modeling the food sources. Each edge $e \in E$ has a positive length $c_{e}$ and a positive capacity $x_{e}(t) ; c_{e}$ is fixed, but $x_{e}(t)$ is a function of time. The resistance $r_{e}(t)$ of $e$ is $r_{e}(t)=c_{e} / x_{e}(t)$. In the electrical network defined by these resistances, a current of value 1 is forced from $s_{0}$ to $s_{1}$. For an (arbitrarily oriented) edge $e=(u, v)$, let $q_{e}(t)$ be the resulting current over $e$. Then, the capacity of $e$ evolves according to the differential equation

$$
\begin{equation*}
\dot{x}_{e}(t)=\left|q_{e}(t)\right|-x_{e}(t), \tag{8.1}
\end{equation*}
$$

where $\dot{x}_{e}$ is the derivative of $x_{e}$ with respect to time. In equilibrium ( $\dot{x}_{e}=0$ for all $e$ ), the flow through any edge is equal to its capacity. In non-equilibrium, the capacity grows (shrinks) if the absolute value of the flow is larger (smaller) than the capacity. In the sequel, we will mostly drop the argument $t$ as is customary in the treatment of dynamical systems. We will also write $q$ for the vector with components $q_{e}$. It is well-known that the electrical flow $q$ is the feasible flow minimizing energy dissipation $\sum_{e} r_{e} q_{e}^{2}$ (Thomson's principle).

We refer to the dynamics above as biologically-grounded, as it was introduced by biologists to model the behavior of a biological system. Miyaji and Ohnishi were the first to analyze convergence for special graphs (parallel links and planar graphs with source and sink on the same face) in [MO08]. In [BMV12] convergence was proven for all graphs. We state the result from [BMV12] for the special case that the shortest path is unique.

Theorem 8.1 ([BMV12]). Assume $x_{e}(0)>0$ and $c_{e}>0$ for all $e$, and that the undirected shortest path $P^{*}$ from $s_{0}$ to $s_{1}$ w.r.t. the cost vector $c$ is unique. Then $x(t)$ in (8.1) converges to $P^{*}$. Namely, $x_{e}(t) \rightarrow 1$ for $e \in P^{*}$ and $x_{e} \rightarrow 0$ for $e \notin P^{*}$ as $t \rightarrow \infty$.
[BMV12] also proves an analogous result for the undirected transportation problem; [Bon13] simplified the argument under additional assumptions. The paper [Bon15] studies a more general dynamics and proves convergence for parallel links.

In this paper, we extend this result to non-negative undirected linear programs

$$
\begin{equation*}
\min \left\{c^{\mathrm{T}} x: A f=b,|f| \leqslant x\right\} \tag{8.2}
\end{equation*}
$$



Figure 8.1: The experiment in [NYT00] (reprinted from there): (a) shows the maze uniformly covered by Physarum; yellow color indicates presence of Physarum. Food (oatmeal) is provided at the locations labeled AG. After a while the mold retracts to the shortest path connecting the food sources as shown in (b) and (c). (d) shows the underlying abstract graph. The video [Phy10] shows the experiment.
where $A \in \mathbb{R}^{n \times m}, b \in \mathbb{R}^{n}, x \in \mathbb{R}^{m}, c \in \mathbb{R}_{\geqslant 0}^{m}$, and the absolute values are taken componentwise. Undirected LPs can model a wide range of problems, e.g., optimization problems such as shortest path and min-cost flow in undirected graphs, and the Basis Pursuit problem in signal processing [CDS98].

We use $n$ for the number of rows of $A$ and $m$ for the number of columns, since this notation is appropriate when $A$ is the node-edge-incidence matrix of a graph. A vector $f \in \mathbb{R}^{m}$ is feasible if $A f=b$. We assume that the system $A f=b$ has a feasible solution and every nonzero vector $f$ in the kernel ${ }^{1}$ of $A$ has positive cost $\sum_{e} c_{e}\left|f_{e}\right|>0$. The vector $q \in \mathbb{R}^{m}$ in (8.1) is now the minimum energy feasible solution

$$
\begin{equation*}
q(t)=\underset{f \in \mathbb{R}^{m}}{\operatorname{argmin}}\left\{\sum_{e: x_{e} \neq 0} \frac{c_{e}}{x_{e}(t)} f_{e}^{2}: A f=b \wedge f_{e}=0 \text { whenever } x_{e}=0\right\} \tag{8.3}
\end{equation*}
$$

We remark that $q$ is unique; see Section 9.2.1. If $A$ is the incidence matrix of a graph (the column corresponding to an edge $e$ has one entry +1 , one entry -1 and all other entries are equal to zero), (8.2) is a transshipment problem with flow sources and sinks encoded by a demand vector $b$. The condition that there is no solution in the kernel of $A$ with $c_{e} f_{e}=0$ for all $e$ states that every cycle contains at least one edge of positive cost. In that setting, $q(t)$ as defined by (8.3) coincides with the electrical flow induced by resistors of value $c_{e} / x_{e}(t)$. We now state our first main result, which is proved in Chapter 9 .

Theorem 8.2. Let $c \geqslant 0$ satisfy $c^{\mathrm{T}}|f|>0$ for every non-zero $f$ in the kernel of $A$. Let $x^{*}$ be an optimum solution of (8.2) and let $X_{\star}$ be the set of optimum solutions. Assume $x(0)>0$. The following holds for the dynamics (8.1) with $q$ as in (8.3):
(i) The solution $x(t)$ exists for all $t \geqslant 0$.
(ii) The cost $c^{\mathrm{T}} x(t)$ converges to $c^{\mathrm{T}} x^{*}$ as $t$ goes to infinity.
(iii) The vector $x(t)$ converges to $X_{\star}$, i.e., $\lim _{t \rightarrow \infty} \inf \left\{\left\|x(t)-x^{\prime}\right\|: x^{\prime} \in X_{\star}\right\} \rightarrow 0$.
(iv) For all $e$ with $c_{e}>0, x_{e}(t)-\left|q_{e}(t)\right|$ converges to zero as $t$ goes to infinity. ${ }^{2}$ If $x^{*}$ is unique, $x(t)$ and $q(t)$ converge to $x^{*}$ as $t$ goes to infinity.

Item (i) was previously shown in [SV16a] for the case of a strictly positive cost vector. The result in [SV16a] is actually stated only for the all-ones cost vector $c=\mathbf{1}$. The case of a general positive cost vector reduces to this special case by rescaling the solution vector $x$. Item $(i)$ for the more general cost vector and items (ii) to (iv) are new. We stress that the dynamics (8.1) is biologically-grounded. It was proposed to model a biological system and not as an optimization method. Nevertheless, it can solve a large class of non-negative LPs. Table 8.1 summarizes our first main result and puts it into context.

Sections 9.1 and 9.2 are devoted to the proof of our first main theorem. For ease of exposition, we present the proof in two steps. In Section 9.1, we give a proof under the following simplifying assumptions:
(A) $c>0$,
(B) The basic feasible solutions of (8.2) have distinct cost,
(C) We start with a positive vector $x(0) \in X_{\text {dom }}:=\left\{x \in \mathbb{R}^{n}\right.$ : there is a feasible $f$ with $\left.|f| \leqslant x\right\}$. Section 9.1 generalizes [Bon13]. For the undirected shortest path problem, condition (B) states that all simple undirected source-sink paths have distinct cost and condition (C) states that all source-sink cuts have a capacity of at least one at time zero (and hence at all times). The existence of a solution with domain $[0, \infty)$ was already shown in [SV16a]. We will show that $X_{\text {dom }}$ is an invariant set, i.e., the solution

[^10]| Reference | Problem <br> (Undirected Case) | Existence <br> of Solution | Convergence <br> to OPT | Comments |
| :---: | :---: | :---: | :---: | :---: |
| $[$ MO08 $]$ | Shortest Path | Yes | Yes | parallel edges, planar graphs |
| $[$ BMV12] | Shortest Path | Yes | Yes | all graphs |
| $[$ SV16a $]$ | Positive LP | Yes | No | $c>0$ |
| Our <br> Result | Nonnegative LP | Yes | Yes | $1) c \geqslant 0$\left\lvert\,$\operatorname{ker}(A): c^{\mathrm{T}}\|v\|>0$ l\right. |

Table 8.1: Convergence results for the continuous undirected Physarum dynamics (8.1).
stays in $X_{\text {dom }}$ for all times, and that $E(x)=\sum_{e} r_{e} x_{e}^{2}=\sum_{e} c_{e} x_{e}$ is a Lyapunov function ${ }^{3}$ [LaS76, Tes12] for the dynamics (8.1), i.e., $\dot{E} \leqslant 0$ and $\dot{E}=0$ if and only if $\dot{x}=0$. It follows from general theorems about dynamical systems that the dynamics converges to a fixed point of (8.1). The fixed points are precisely the vectors $|f|$ where $f$ is a feasible solution of (8.2). A final argument establishes that the dynamics converges to a fixed point of minimum cost.

In Section 9.2, we prove the general case of the first main theorem. We assume
(D) $c \geqslant 0$,
(E) $\operatorname{cost}(z)=c^{\mathrm{T}}|z|>0$ for every non-zero vector $z$ in the kernel of $A$,
(F) We start with a positive vector $x(0)>0$.

Section 9.2 generalizes [BMV12] in two directions. First, we treat general undirected LPs and not just the undirected shortest path problem, respectively, the transshipment problem. Second, we replace the condition $c>0$ by the requirement $c \geqslant 0$ and every non-zero vector in the kernel of $A$ has positive cost. For the undirected shortest path problem, the latter condition states that the underlying undirected graph has no zero-cost cycle. Section 9.2 is technically considerably more difficult than Section 9.1. We first establish the existence of a solution with domain $[0, \infty)$. To this end, we derive a closed formula for the minimum energy feasible solution and prove that the mapping $x \mapsto q$ is locally Lipschitz. Existence of a solution with domain $[0, \infty)$ follows by standard arguments. We then show that $X_{\text {dom }}$ is an attractor, i.e., the solution $x(t)$ converges to $X_{\text {dom }}$. We next characterize equilibrium points and exhibit a Lyapunov function. The Lyapunov function is a normalized version of $E(x)$. The normalization factor is equal to the optimal value of the linear program $\max \{\alpha: A f=\alpha b,|f| \leqslant x\}$ in the variables $f$ and $\alpha$. Convergence to an equilibrium point follows from the existence of a Lyapunov function. A final argument establishes that the dynamics converges to a fixed point of minimum cost.

### 8.2 The Biologically-Inspired Model

Ito et al. [IJNT11] initiated the study of the dynamics

$$
\begin{equation*}
\dot{x}(t)=q(t)-x(t) \tag{8.4}
\end{equation*}
$$

We refer to this dynamics as the directed dynamics in contrast to the undirected dynamics (8.1). The directed dynamics is biologically-inspired - the similarity to (8.1) is the inspiration. It was never claimed to model the behavior of a biological system. Rather, it was introduced as a biologically-inspired optimization method. The work in [IJNT11] shows convergence of this directed dynamics (8.4) for the directed shortest path problem and [JZ12, SV16c, Bon16] show convergence for general positive linear programs, i.e., linear programs with positive cost vector $c>0$ of the form

$$
\begin{equation*}
\min \left\{c^{\mathrm{T}} x: A x=b, x \geqslant 0\right\} \tag{8.5}
\end{equation*}
$$

The discrete versions of both dynamics define sequences $x^{(t)}, t=0,1,2, \ldots$ through

$$
\begin{array}{lr}
x^{(t+1)}=\left(1-h^{(t)}\right) x^{(t)}+h^{(t)} q^{(t)} & \text { discrete directed dynamics; } \\
x^{(t+1)}=\left(1-h^{(t)}\right) x^{(t)}+h^{(t)}\left|q^{(t)}\right| & \text { discrete undirected dynamics } \tag{8.7}
\end{array}
$$

where $h^{(t)}$ is the step size and $q^{(t)}$ is the minimum energy feasible solution as in (8.3). For the discrete dynamics, we can ask complexity questions. This is particularly relevant for the discrete directed dynamics as it was designed as an biologically-inspired optimization method.

[^11]For completeness, we review the state-of-the-art results for the discrete undirected dynamics. For the undirected shortest path problem, the convergence of the discrete undirected dynamics (8.7) was shown in $\left[\mathrm{BBD}^{+} 13\right]$. The convergence proof gives an upper bound on the step size and on the number of steps required until an $\varepsilon$-approximation of the optimum is obtained. [SV16b] extends the result to the transshipment problem and [SV16a] further generalizes the result to the case of positive LPs. The paper [SV16b] is related to our first result. It shows convergence of the discretized undirected dynamics (8.7), we show convergence of the continuous undirected dynamics (8.1) for a more general cost vector.

We come to the discrete directed Physarum-inspired dynamics (8.6). Similarly to the undirected setting, Becchetti et al. $\left[\mathrm{BBD}^{+} 13\right]$ showed the convergence of (8.6) for the shortest path problem. Straszak and Vishnoi extended the analysis to the transshipment problem [SV16b] and positive LPs [SV16c].

Theorem 8.3. [SV16c, Theorem 1.3] Let $A \in \mathbb{Z}^{n \times m}$ have full row rank $(n \leqslant m), b \in \mathbb{Z}^{n}, c \in \mathbb{Z}_{>0}^{m}$, and let $D_{S} \stackrel{\text { def }}{=} \max \{|\operatorname{det}(M)|: M$ is a square sub-matrix of $A\} .{ }^{4}$ Suppose the Physarum-inspired dynamics (8.6) is initialized with a feasible point $x^{(0)}$ of (8.5) such that $M^{-1} \leqslant x^{(0)} \leqslant M$ and $c^{\mathrm{T}} x^{(0)} \leqslant M \cdot$ opt for some $M \geqslant 1$, where opt denotes the optimum cost of (8.5). Then, for any $\varepsilon>0$ and step size $h \leqslant \varepsilon /\left(\sqrt{6}\|c\|_{1} D_{S}\right)^{2}$, after $k=O\left((\varepsilon h)^{-2} \ln M\right)$ steps, $x^{(k)}$ is a feasible solution with $c^{\mathrm{T}} x^{(k)} \leqslant(1+\varepsilon) \mathrm{opt}$.

Theorem 8.3 gives an algorithm that computes a $(1+\varepsilon)$-approximation to the optimal cost of (8.5). In comparison to $\left[\mathrm{BBD}^{+} 13, \mathrm{SV} 16 \mathrm{~b}\right]$, it has several shortcomings. First, it requires a feasible starting point. Second, the step size depends linearly on $\varepsilon$. Third, the number of steps required to reach an $\varepsilon$-approximation has a quartic dependence on opt $/(\varepsilon \Phi)$. In contrast, the analysis in [ $\mathrm{BBD}^{+} 13$, SV16b] yields a step size independent of $\varepsilon$ and a number of steps that depends only logarithmically on $1 / \varepsilon$, see Table 8.2.

We overcome these shortcomings in Chapter 10. Before we can state our result, we need some notation. Let $X_{\star}$ be the set of optimal solutions to (8.5). The distance of a capacity vector $x$ to $X_{\star}$ is defined as

$$
\operatorname{dist}\left(x, X_{\star}\right) \stackrel{\text { def }}{=} \inf \left\{\left\|x-x^{\prime}\right\|_{\infty}: x^{\prime} \in X_{\star}\right\}
$$

Let $\gamma_{A} \stackrel{\text { def }}{=} \operatorname{gcd}\left(\left\{A_{i j}: A_{i j} \neq 0\right\}\right) \in \mathbb{Z}_{>0}$ and

$$
\begin{equation*}
D \stackrel{\text { def }}{=} \max \left\{|\operatorname{det}(M)|: M \text { is a square submatrix of } A / \gamma_{A} \text { with dimension } n-1 \text { or } n\right\} \tag{8.8}
\end{equation*}
$$

Let $\mathcal{N}$ be the set of non-optimal basic feasible solutions of (8.5) and

$$
\begin{equation*}
\Phi \stackrel{\text { def }}{=} \min _{g \in \mathcal{N}} c^{\mathrm{T}} g-\mathrm{opt} \geqslant 1 /\left(D \gamma_{A}\right)^{2} \tag{8.9}
\end{equation*}
$$

where the inequality is a well result in combinatorial optimization [PS82, Lemma 8.6]. For completeness, we present a proof in Section 10.5. Informally, our second main result proves the following properties of the Physarum-inspired dynamics (8.6):
(i) For any $\varepsilon>0$ and any strongly dominating starting point ${ }^{5} x^{(0)}$, there is a fixed step size $h\left(x^{(0)}\right)$ such that the Physarum-inspired dynamics (8.6) initialized with $x^{(0)}$ and $h\left(x^{(0)}\right)$ converges to $X_{\star}$, i.e., $\operatorname{dist}\left(x^{(k)}, X_{\star}\right)<\varepsilon /\left(D \gamma_{A}\right)$ for large enough $k$.
(ii) The step size can be chosen independently of $\varepsilon$.
(iii) The number of steps $k$ depends logarithmically on $1 / \varepsilon$ and quadratically on opt/ $\Phi$.
(iv) The efficiency bounds depend on the scale-invariant determinant ${ }^{6} D$.

In Section 10.8, we establish a corresponding lower bound. We show that for the Physarum-inspired dynamics (8.6) to compute a point $x^{(k)}$ such that $\operatorname{dist}\left(x^{(k)}, X_{\star}\right)<\varepsilon$, the number of steps required for computing an $\varepsilon$-approximation has to grow linearly in opt $/(h \Phi)$ and $\ln (1 / \varepsilon)$, i.e. $k \geqslant \Omega\left(\right.$ opt $\cdot(h \Phi)^{-1}$. $\ln (1 / \varepsilon))$. Table 8.2 puts our results into context.

[^12]| Reference | Problem <br> (Directed Case) | step size $h$ | number of steps $k$ | Guarantee |
| :---: | :---: | :---: | :---: | :---: |
| $\left[\mathrm{BBD}^{+} 13\right]$ | Shortest Path | indep. of $\varepsilon$ | $\operatorname{poly}\left(m, n,\\|c\\|_{1},\left\\|x^{(0)}\right\\|_{1}\right)$ <br> $\cdot \ln (1 / \varepsilon)$ | $\operatorname{dist}\left(x^{(k)}, X_{\star}\right)<\varepsilon$ |
| $[\mathrm{SV16b}]$ | Transshipment | indep. of $\varepsilon$ | $\operatorname{poly}\left(m, n,\\|c\\|_{1},\\|b\\|_{1},\left\\|x^{(0)}\right\\|_{1}\right)$ <br> $\cdot \ln (1 / \varepsilon)$ | $\operatorname{dist}\left(x^{(k)}, X_{\star}\right)<\varepsilon$ |
| $[\mathrm{SV16c}]$ | Positive LP | depends on $\varepsilon$ | $\operatorname{poly}\left(\\|c\\|_{1}, D_{S}, \ln \left\\|x^{(0)}\right\\|_{1}\right)$ <br> $1 /(\Phi \varepsilon)^{4}$ | $c^{\mathrm{T}} x^{(k)} \leqslant(1+\varepsilon) \mathrm{opt}$ <br> $c^{\mathrm{T}} x^{(k)}<\min _{g \in \mathcal{N}} c^{\mathrm{T}} g$ |
| Our <br> Result | Positive LP | indep. of $\varepsilon$ | $\operatorname{poly}\left(\\|c\\|_{1},\\|b\\|_{1}, D, \gamma_{A}, \ln \left\\|x^{(0)}\right\\|_{1}\right)$ <br> $\Phi^{-2} \ln (1 / \varepsilon)$ | $\operatorname{dist(x^{(k)},X_{\star })<\frac {\varepsilon }{D\gamma _{A}}}$ |
| Lower <br> Bound | Positive LP | indep. of $\varepsilon$ | $\Omega\left(\right.$ opt $\left.\cdot(h \Phi)^{-1} \ln (1 / \varepsilon)\right)$ | $\operatorname{dist(x^{(k)},X_{\star })<\varepsilon }$ |

Table 8.2: Convergence results for the discrete directed Physarum-inspired dynamics (8.6).

We state now our second main result for the special case of a feasible starting point, and we provide the full version in Theorem 10.2 which applies for arbitrary strongly dominating starting point, see Section 10.1. We use the following constants in the statement of the bounds.
(i) $h_{0} \stackrel{\text { def }}{=} c_{\text {min }} /\left(4 D\|c\|_{1}\right)$, where $c_{\text {min }} \stackrel{\text { def }}{=} \min _{i}\left\{c_{i}\right\}$;
(ii) $\Psi^{(0)} \stackrel{\text { def }}{=} \max \left\{m D^{2}\left\|b / \gamma_{A}\right\|_{1},\left\|x^{(0)}\right\|_{\infty}\right\}$;
(iii) $C_{1} \stackrel{\text { def }}{=} D\left\|b / \gamma_{A}\right\|_{1}\|c\|_{1}, C_{2} \stackrel{\text { def }}{=} 8^{2} m^{2} n D^{5} \gamma_{A}^{2}\|A\|_{\infty}\|b\|_{1}$ and $C_{3} \stackrel{\text { def }}{=} D^{3} \gamma_{A}\|b\|_{1}\|c\|_{1}$.

Theorem 8.4. Suppose $A \in \mathbb{Z}^{n \times m}$ has full row $\operatorname{rank}(n \leqslant m), b \in \mathbb{Z}^{n}, c \in \mathbb{Z}_{>0}^{m}$ and $\varepsilon \in(0,1)$. Given a feasible starting point $x^{(0)}>0$ the Physarum-inspired dynamics (8.6) with step size $h \leqslant(\Phi / \mathrm{opt}) \cdot h_{0}^{2} / 2$ outputs for any $k \geqslant 4 C_{1} /(h \Phi) \cdot \ln \left(C_{2} \Psi^{(0)} /\left(\varepsilon \cdot \min \left\{1, x_{\min }^{(0)}\right\}\right)\right)$ a feasible $x^{(k)}>0$ such that $\operatorname{dist}\left(x^{(k)}, X_{\star}\right)<$ $\varepsilon /\left(D \gamma_{A}\right)$.

We stated the bounds on $h$ in terms of the unknown quantities $\Phi$ and opt. However, $\Phi /$ opt $\geqslant 1 / C_{3}$ by Lemma 9.10 and hence replacing $\Phi /$ opt by $1 / C_{3}$ yields constructive bounds for $h$. Note that the upper bound on the step size does not depend on $\varepsilon$ and that the bound on the number of iterations depends logarithmically on $1 / \varepsilon$ and quadratically on opt $/ \Phi$.

What can be done if the initial point is not strongly dominating? For the transshipment problem it suffices to add an edge of high capacity and high cost from every source node to every sink node [ $\mathrm{BBD}^{+} 13$, SV16b]. This will make the instance strongly dominating and will not affect the optimal solution. We generalize this observation to positive linear programs. We add an additional column equal to $b$ and give it sufficiently high capacity and cost. This guarantees that the resulting instance is strongly dominating and the optimal solution remains unaffected. Moreover, our approach generalizes and improves upon [SV16b, Theorem 1.2], see Section 10.7.

Proof Techniques: The crux of the analysis in [IJNT11, $\mathrm{BBD}^{+} 13, \mathrm{SV} 16 \mathrm{~b}$ ] is to show that for large enough $k, x^{(k)}$ is close to a non-negative flow $f^{(k)}$ and then to argue that $f^{(k)}$ is close to an optimal flow $f^{\star}$. This line of arguments yields a convergence of $x^{(k)}$ to $X_{\star}$ with a step size $h$ chosen independently of $\varepsilon$.

In Chapter 10, we extend the preceding approach to positive linear programs, by generalizing the concept of non-negative cycle-free flows to non-negative feasible kernel-free vectors (Section 10.4). Although, we use the same high level ideas as in $\left[\mathrm{BBD}^{+} 13, \mathrm{SV} 16 \mathrm{~b}\right]$, we stress that our analysis generalizes all relevant lemmas in $\left[\mathrm{BBD}^{+} 13, \mathrm{SV} 16 \mathrm{~b}\right]$ and it uses arguments from linear algebra and linear programming duality, instead of combinatorial arguments. Further, our core efficiency bounds (Section 10.2) extend [SV16c] and yield a scale-invariant determinant dependence of the step size and are applicable for any strongly dominating starting point (Section 10.3).

## Chapter 9

## Biologically-Grounded Physarum Dynamics

### 9.1 Convergence: Simple Instances

In this section, we prove Theorem 8.2 under the simplifying assumptions (A) to (C), defined in page 88 .

### 9.1.1 Preliminaries

Note that we may assume that $A$ has full row-rank since any equation that is linearly dependent on other equations can be deleted without changing the feasible set. We continue to use $n$ and $m$ for the dimension of $A$. Thus, $A$ has rank $n$. We continue by fixing some terms and notation. A basic feasible solution of (8.2) is a pair of vectors $x$ and $f=\left(f_{B}, f_{N}\right)$, where $f_{B}=A_{B}^{-1} b$ and $A_{B}$ is a square $n \times n$ non-singular sub-matrix of $A$ and $f_{N}=0$ is the vector indexed by the coordinates not in $B$, and $x=|f|$. Since $f$ uniquely determines $x$, we may drop the latter for the sake of brevity and call $f$ a basic feasible solution of (8.2). A feasible solution $f$ is kernel-free or non-circulatory if it is contained in the convex hull of the basic feasible solutions. ${ }^{1}$ We say that a vector $f^{\prime}$ is sign-compatible with a vector $f$ (of the same dimension) or $f$-sign-compatible if $f_{e}^{\prime} \neq 0$ implies $f_{e}^{\prime} f_{e}>0$. In particular, $\operatorname{supp}\left(f^{\prime}\right) \subseteq \operatorname{supp}(f)$. For a given capacity vector $x$ and a vector $f \in \mathbb{R}^{m}$ with $\operatorname{supp}(f) \subseteq \operatorname{supp}(x)$, we use $E(f)=\sum_{e}\left(c_{e} / x_{e}\right) f_{e}^{2}$ to denote the energy of $f$. The energy of $f$ is infinite, if $\operatorname{supp}(f) \nsubseteq \operatorname{supp}(x)$. We use $\operatorname{cost}(f)=\sum_{e} c_{e}\left|f_{e}\right|=c^{\mathrm{T}}|f|$ to denote the cost of $f$. Note that $E(x)=\sum_{e}\left(c_{e} / x_{e}\right) x_{e}^{2}=\sum_{e} c_{e} x_{e}=\operatorname{cost}(x)$. We define the constants $c_{\text {max }}=\|c\|_{\infty}$ and $c_{\text {min }}=\min _{e: c_{e}>0} c_{e}$.

We use the following corollary of the finite basis theorem for polyhedra.
Lemma 9.1. Let $f$ be a feasible solution of (8.2). Then $f$ is the sum of a convex combination of at most $m$ basic feasible solutions plus a vector in the kernel of $A$. Moreover, all elements in this representation are sign-compatible with $f$.

Proof. We may assume $f \geqslant 0$. Otherwise, we flip the sign of the appropriate columns of $A$. Thus, the system $A f=b, f \geqslant 0$ is feasible and $f$ is the sum of a convex combination of at most $m$ basic feasible solutions plus a vector in the kernel of $A$ by the finite basis theorem [Sch99, Corollary 7.1b]. By definition, the elements in this representation are non-negative vectors and hence sign-compatible with $f$.

Lemma 9.2 (Grönwall's Lemma). Let $A, B, \alpha, \beta \in \mathbb{R}, \alpha \neq 0, \beta \neq 0$, and let $g$ be a continuous differentiable function on $[0, \infty)$. If $A+\alpha g(t) \leqslant \dot{g}(t) \leqslant B+\beta g(t)$ for all $t \geqslant 0$, then $-A / \alpha+(g(0)+A / \alpha) e^{\alpha t} \leqslant$ $g(t) \leqslant-B / \beta+(g(0)+B / \beta) e^{\beta t}$ for all $t \geqslant 0$.
Proof. We show the upper bound. Assume first that $B=0$. Then

$$
\frac{d}{d t} \frac{g}{e^{\beta t}}=\frac{\dot{g} e^{\beta t}-\beta g e^{\beta t}}{e^{2 \beta t}} \leqslant 0 \quad \text { implies } \quad \frac{g(t)}{e^{\beta t}} \leqslant \frac{g(0)}{e^{\beta 0}}=g(0) .
$$

If $B \neq 0$, define $h(t)=g(t)+B / \beta$. Then

$$
\dot{h}=\dot{g} \leqslant B+\beta g=B+\beta(h-B / \beta)=\beta h
$$

and hence $h(t) \leqslant h(0) e^{\beta t}$. Therefore $g(t) \leqslant-B / \beta+(g(0)+B / \beta) e^{\beta t}$.
An immediate consequence of Grönwall's Lemma is that the undirected Physarum dynamics (8.1) initialized with any positive starting vector $x(0)$, generates a trajectory $\{x(t)\}_{t \geqslant 0}$ such that each time state $x(t)$ is a positive vector. Indeed, since $\dot{x}_{e}=\left|q_{e}\right|-x_{e} \geqslant-x_{e}$, we have $x_{e}(t) \geqslant x_{e}(0) \cdot \exp \{-t\}$ for every index $e$ with $x_{e}(0)>0$ and every time $t$. Further, by (8.1) and (8.3), it holds for indices $e$ with $x_{e}(0)=0$ that $x_{e}(t)=0$ for every time $t$. Hence, the trajectory $\{x(t)\}_{t \geqslant 0}$ has a time-invariant support.

[^13]Lemma 9.3 ([JZ12]). Let $R=\operatorname{diag}\left(c_{e} / x_{e}\right)$. Then $q=R^{-1} A^{\mathrm{T}} p$, where $p=\left(A R^{-1} A^{\mathrm{T}}\right)^{-1} b$.
Proof. $q$ minimizes $\sum_{e} r_{e} q_{e}^{2}$ subject to $A q=b$. The Karush-Kuhn-Tucker (KKT) optimality conditions for constrained optimization [Boy04] imply the existence of a vector $p$ such that $R q=A^{\mathrm{T}} p$. Substituting into $A q=b$ yields $p=\left(A R^{-1} A^{\mathrm{T}}\right)^{-1} b$.

Lemma 9.4. $X_{\mathrm{dom}}$ is an invariant set, i.e., if $x(0) \in X_{\mathrm{dom}}$ then $x(t) \in X_{\mathrm{dom}}$ for all $t$.
Proof. Let $q(t)$ be the minimum energy feasible solution with respect to $R(t)=\operatorname{diag}\left(c_{e} / x_{e}(t)\right)$, and let $f(t)$ be such that $f(0)$ is feasible, $|f(0)| \leqslant x(0)$, and $\dot{f}(t)=q(t)-f(t)$. Then $\frac{d}{d t}(A f-b)=A(q-f)=b-A f$ and hence $A f(t)-b=(A f(0)-b) e^{-t}=0$. Thus $f(t)$ is feasible for all $t$. Moreover,

$$
\frac{d}{d t}(f-x)=\dot{f}-\dot{x}=q-f-(|q|-x)=q-|q|-(f-x) \leqslant-(f-x)
$$

Thus $f(t)-x(t) \leqslant(f(0)-x(0)) e^{-t} \leqslant 0$ by Grönwall's Lemma applied with $g(t)=f(t)-x(t)$ and $\beta=-1$, and hence $f(t) \leqslant x(t)$ for all $t$. Similarly,

$$
\frac{d}{d t}(f+x)=\dot{f}+\dot{x}=q-f+(|q|-x)=q+|q|-(f+x) \geqslant-(f+x)
$$

Thus $f(t)+x(t) \geqslant(f(0)+x(0)) e^{-t} \geqslant 0$ by Grönwall's Lemma applied with $g(t)=f(t)+x(t)$ and $\alpha=-1$ and $A=0$, and hence $f(t) \geqslant-x(t)$ for all $t$.

We conclude that $|f(t)| \leqslant x(t)$ for all $t$. Thus, $x(t) \in X_{\text {dom }}$ for all $t$.

### 9.1.2 The Convergence Proof

We will first characterize the equilibrium points. They are precisely the points $|f|$, where $f$ is a basic feasible solution; the proof uses assumption (B) in page 88 . We then show that $E(x)$ is a Lyapunov function for (8.1), in particular, $\dot{E} \leqslant 0$ and $\dot{E}=0$ if and only if $x$ is an equilibrium point. For this argument, we need that the energy of $q$ is at most the energy of $x$ with equality if and only if $x$ is an equilibrium point. This proof uses assumptions (A) and (C) in page 88. It follows from the general theory of dynamical systems that $x(t)$ approaches an equilibrium point. Finally, we show that convergence to a non-optimal equilibrium is impossible.

Lemma 9.5 (Generalization of Lemma 2.3 in [Bon13]). Assume (A) to (C). If $f$ is a basic feasible solution of (8.2), then $x=|f|$ is an equilibrium point. Conversely, if $x$ is an equilibrium point, then $x=|f|$ for some basic feasible solution $f$.

Proof. Let $f$ be a basic feasible solution, let $x=|f|$, and let $q$ be the minimum energy feasible solution with respect to the resistances $c_{e} / x_{e}$. We have $A q=b$ and $\operatorname{supp}(q) \subseteq \operatorname{supp}(x)$ by definition of $q$. Since $f$ is a basic feasible solution there is a subset $B$ of size $n$ of the columns of $A$ such that $A_{B}$ is nonsingular and $f=\left(A_{B}^{-1} b, 0\right)$. Since $\operatorname{supp}(q) \subseteq \operatorname{supp}(x) \subseteq B$, we have $q=\left(q_{B}, 0\right)$ for some vector $q_{B}$. Thus, $b=A q=A_{B} q_{B}$ and hence $q_{B}=f_{B}$. Therefore $\dot{x}=|q|-x=0$ and $x$ is an equilibrium point.

Conversely, if $x$ is an equilibrium point, $\left|q_{e}\right|=x_{e}$ for every $e$. By changing the signs of some columns of $A$, we may assume $q \geqslant 0$. Then $q=x$. Since $q_{e}=x_{e} / c_{e} A_{e}^{\mathrm{T}} p$ where $A_{e}$ is the $e$-th column of $A$ by Lemma 9.3, we have $c_{e}=A_{e}^{\mathrm{T}} p$, whenever $x_{e}>0$. By Lemma 9.1, $q$ is a convex combination of basic feasible solutions and a vector in the kernel of $A$ that are sign-compatible with $q$. The vector in the kernel must be zero as $q$ is a minimum energy feasible solution. For any basic feasible solution $z$ contributing to $q$, we have $\operatorname{supp}(z) \subseteq \operatorname{supp}(x)$. Summing over the $e \in \operatorname{supp}(z)$, we obtain $\operatorname{cost}(z)=\sum_{e \in \operatorname{supp}(z)} c_{e} z_{e}=$ $\sum_{e \in \operatorname{supp}(z)} z_{e} A_{e}^{\mathrm{T}} p=b^{\mathrm{T}} p$. Thus, the convex combination involves only a single basic feasible solution by assumption (B) and hence $x$ is a basic feasible solution.

The vector $x(t)$ dominates a feasible solution at all times. Since $q(t)$ is the minimum energy feasible solution at time $t$, this implies $E(q(t)) \leqslant E(x(t))$ at all times. A further argument shows that we have equality if and only if $x=|q|$.

Lemma 9.6 (Generalization of Lemma 3.1 in [Bon13]). Assume (A) to (C). At all times, $E(q) \leqslant E(x)$. If $E(q)=E(x)$, then $x=|q|$.

Proof. Recall that $x(t) \in X_{\text {dom }}$ for all $t$. Thus, at all times, there is a feasible $f$ such that $|f| \leqslant x$. Since $q$ is a minimum energy feasible solution, we have

$$
E(q) \leqslant E(f) \leqslant E(x)
$$

If $E(q)=E(x)$ then $E(q)=E(f)$ and hence $q=f$ since the minimum energy feasible solution is unique. Also, $|f|=x$ since $|f| \leqslant x$ and $\left|f_{e}\right|<x_{e}$ for some $e$ implies $E(f)<E(x)$. The last conclusion uses $c>0$.

Lyapunov functions are the main tool for proving convergence of dynamical systems. We show that $E(x)$ is a Lyapunov function for (8.1).

Lemma 9.7 (Generalization of Lemma 3.2 in [Bon13]). Assume (A) to (C). $E(x)$ is a Lyapunov function for (8.1), i.e., it is continuous as a function of $x, E(x) \geqslant 0, \dot{E}(x) \leqslant 0$ and $\dot{E}(x)=0$ if and only if $\dot{x}=0$.

Proof. $E$ is clearly continuous and non-negative. Recall that $E(x)=\operatorname{cost}(x)$. Let $R$ be the diagonal matrix with entries $c_{e} / x_{e}$. Then

$$
\begin{aligned}
\frac{d}{d t} \operatorname{cost}(x) & =c^{\mathrm{T}}(|q|-x) \\
& =x^{\mathrm{T}} R|q|-x^{\mathrm{T}} R x \\
& =x^{\mathrm{T}} R^{1 / 2} R^{1 / 2}|q|-x^{\mathrm{T}} R x \\
& \leqslant\left(q^{\mathrm{T}} R q\right)^{1 / 2}\left(x^{\mathrm{T}} R x\right)^{1 / 2}-x^{\mathrm{T}} R x \\
& \leqslant\left(x^{\mathrm{T}} R x\right)^{1 / 2}\left(x^{\mathrm{T}} R x\right)^{1 / 2}-x^{\mathrm{T}} R x \\
& =0
\end{aligned}
$$

Observe that $\frac{d}{d t} \operatorname{cost}(x)=0$ implies that both inequalities above are equalities. This is only possible if the vectors $|q|$ and $x$ are parallel and $E(q)=E(x)$. Thus, $x=|q|$ by Lemma 9.6.

It follows now from the general theory of dynamical systems that $x(t)$ converges to an equilibrium point.

Corollary 9.8 (Generalization of Corollary 3.3. in [Bon13].). Assume (A) to (C). As $t \rightarrow \infty, x(t)$ approaches an equilibrium point and $c^{\mathrm{T}} x(t)$ approaches the cost of the corresponding basic feasible solution.

Proof. The proof in [Bon13] carries over. We include it for completeness. The existence of a Lyapunov function $E$ implies by [LaS76, Corollary 2.6.5] that $x(t)$ approaches the set $\left\{x \in \mathbb{R}_{\geqslant 0}^{m}: \dot{E}=0\right\}$, which by Lemma 9.7 is the same as the set $\left\{x \in \mathbb{R}_{\geqslant 0}^{m}: \dot{x}=0\right\}$. Since this set consists of isolated points (Lemma 9.5), $x(t)$ must approach one of those points, say the point $x_{0}$. When $x=x_{0}$, one has $E(q)=$ $E(x)=\operatorname{cost}(x)=c^{\mathrm{T}} x$.

It remains to exclude that $x(t)$ converges to a non-optimal equilibrium point.
Theorem 9.9 (Generalization of Theorem 3.4 in [Bon13]). Assume (A) to (C). As $t \rightarrow \infty, c^{\mathrm{T}} x(t)$ converges to the cost of the optimal solution and $x(t)$ converges to the optimal solution.

Proof. By the corollary, it suffices to prove the second part of the claim. For the second part, assume that $x(t)$ converges to a non-optimal solution $z$. Let $x^{*}$ be the optimal solution and let $W=\sum_{e} x_{e}^{*} c_{e} \ln x_{e}$. Let $\delta=\left(\operatorname{cost}(z)-\operatorname{cost}\left(x^{*}\right)\right) / 2$. Note that for all sufficiently large $t$, we have $E(q(t)) \geqslant \operatorname{cost}(z)-\delta \geqslant \operatorname{cost}\left(x^{*}\right)+\delta$. Further, by definition $q_{e}=\left(x_{e} / c_{e}\right) A_{e}^{\mathrm{T}} p$ and thus

$$
\dot{W}=\sum_{e} x_{e}^{*} c_{e} \frac{\left|q_{e}\right|-x_{e}}{x_{e}}=\sum_{e} x_{e}^{*}\left|A_{e}^{\mathrm{T}} p\right|-\operatorname{cost}\left(x^{*}\right) \geqslant \sum_{e} x_{e}^{*} A_{e}^{\mathrm{T}} p-\operatorname{cost}\left(x^{*}\right) \geqslant \delta,
$$

where the last inequality follows by $\sum_{e} x_{e}^{*} A_{e}^{\mathrm{T}} p=b^{\mathrm{T}} p=E(q) \geqslant \operatorname{cost}\left(x^{*}\right)+\delta$. Hence $W \rightarrow \infty$, a contradiction to the fact that $x$ is bounded.

### 9.2 Convergence: General Instances

In this section, we prove Theorem 8.2 under the more general assumptions (D) to (F), defined in page 89 .

### 9.2.1 Existence of a Solution with Domain $[0, \infty)$

In this subsection we show that a solution $x(t)$ to (8.1) has domain $[0, \infty)$. We first derive an explicit formula for the minimum energy feasible solution $q$ and then show that the mapping $x \mapsto q$ is Lipschitz continuous; this implies existence of a solution with domain $[0, \infty)$ by standard arguments.

## The Minimum Energy Solution

Recall that for $\gamma_{A}=\operatorname{gcd}\left(\left\{A_{i j}: A_{i j} \neq 0\right\}\right) \in \mathbb{Z}_{>0}$, we defined by

$$
D=\max \left\{|\operatorname{det}(M)|: M \text { is a square submatrix of } A / \gamma_{A} \text { with dimension } n-1 \text { or } n\right\} .
$$

We derive now properties of the minimum energy solution. In particular, if every non-zero vector in the kernel of $A$ has positive cost,
(i) the minimum energy feasible solution is kernel-free and unique (Lemma 9.11),
(ii) $\left|q_{e}\right| \leqslant D\left\|b / \gamma_{A}\right\|_{1}$ for every $e \in[m]$ (Lemma 9.12),
(iii) $q$ is defined by (9.3) (Lemma 9.13), and
(iv) $E(q)=b^{\mathrm{T}} p$, where $p$ is defined by (9.3) (Lemma 9.14).

We note that for positive cost vector $c>0$, these results are known.
We proceed by establishing some useful properties on basic feasible solutions.
Lemma 9.10. Suppose $A \in \mathbb{Z}^{n \times m}$ is an integral matrix, and $b \in \mathbb{Z}^{n}$ is an integral vector. Then, for any basic feasible solutions $f$ with $A f=b$ and $f \geqslant 0$, it holds that $\|f\|_{\infty} \leqslant D\left\|b / \gamma_{A}\right\|_{1}$ and $f_{j} \neq 0$ implies $\left|f_{j}\right| \geqslant 1 /\left(D \gamma_{A}\right)$.

Proof. Since $f$ is a basic feasible solution, it has the form $f=\left(f_{B}, 0\right)$ such that $A_{B} \cdot f_{B}=b$ where $A_{B} \in \mathbb{Z}^{n \times n}$ is an invertible submatrix of $A$. We write $M_{-i,-j}$ to denote the matrix $M$ with deleted $i$-th row and $j$-th column. Let $Q_{j}$ be the matrix formed by replacing the $j$-th column of $A_{B}$ by the column vector $b$. Then, using the fact that $\operatorname{det}(t A)=t^{n} \operatorname{det}(A)$ for every $A \in \mathbb{R}^{n \times n}$ and $t \in \mathbb{Z}$, Cramer's rule yields

$$
\left|f_{B}(j)\right|=\left|\frac{\operatorname{det}\left(Q_{j}\right)}{\operatorname{det}\left(A_{B}\right)}\right|=\frac{1}{\gamma_{A}}\left|\sum_{k=1}^{n} \frac{(-1)^{j+k} \cdot b_{k} \cdot \operatorname{det}\left(\gamma_{A}^{-1}\left[A_{B}\right]_{-k,-j}\right)}{\operatorname{det}\left(\gamma_{A}^{-1} A_{B}\right)}\right|
$$

By the choice of $\gamma_{A}$, the values $\operatorname{det}\left(\gamma_{A}^{-1} A_{B}\right)$ and $\operatorname{det}\left(\gamma_{A}^{-1}\left[A_{B}\right]_{-k,-j}\right)$ are integral for all $k$, it follows that

$$
\left|f_{B}(j)\right| \leqslant D\left\|b / \gamma_{A}\right\|_{1} \quad \text { and } \quad f_{B}(j) \neq 0 \quad \Longrightarrow \quad \frac{1}{D \gamma_{A}} \leqslant\left|f_{B}(j)\right|
$$

Lemma 9.11. If every non-zero vector in the kernel of $A$ has positive cost, the minimum energy feasible solution is kernel-free and unique.

Proof. Let $q$ be a minimum energy feasible solution. Since $q$ is feasible, it can be written as $q_{n}+q_{r}$, where $q_{n}$ is a convex combination of basic feasible solutions and $q_{r}$ lies in the kernel of $A$. Moreover, all elements in this representation are sign-compatible with $q$ by Lemma 9.1. If $q_{r} \neq 0$, the vector $q-q_{r}$ is feasible and has smaller energy, a contradiction. Thus $q_{r}=0$.

We next prove uniqueness. Assume for the sake of a contradiction that there are two distinct minimum energy feasible solutions $q^{(1)}$ and $q^{(2)}$. We show that the solution $\left(q^{(1)}+q^{(2)}\right) / 2$ uses less energy than $q^{(1)}$ and $q^{(2)}$. Since $h \mapsto h^{2}$ is a strictly convex function from $\mathbb{R}$ to $\mathbb{R}$, the average of the two solutions will be better than either solution if there is an index $e$ with $r_{e}>0$ and $q_{e}^{(1)} \neq q_{e}^{(2)}$. The difference $z=q^{(1)}-q^{(2)}$ lies in the kernel of $A$ and hence $\operatorname{cost}(z)=\sum_{e} c_{e}\left|z_{e}\right|>0$. Thus there is an $e$ with $c_{e}>0$ and $z_{e} \neq 0$. We have now shown uniqueness.

Lemma 9.12. Assume that every non-zero vector in the kernel of $A$ has positive cost. Let $q$ be the minimum energy feasible solution. Then $\left|q_{e}\right| \leqslant D\left\|b / \gamma_{A}\right\|_{1}$ for every e.

Proof. Since $q$ is a convex combination of basic feasible solutions, $\left|q_{e}\right| \leqslant \max _{z}\left|z_{e}\right|$ where $z$ ranges over basic feasible solutions of the form $\left(z_{B}, 0\right)$, where $z_{B}=A_{B}^{-1} b$ and $A_{B} \in \mathbb{R}^{n \times n}$ is a non-singular submatrix of $A$. Thus, by Lemma 9.10 every component of $z$ is bounded by $D\left\|b / \gamma_{A}\right\|_{1}$.

In [SV16c], the bound $\left|q_{e}\right| \leqslant D^{2} m\|b\|_{1}$ was shown. We will now derive explicit formulae for the minimum energy solution $q$. We will express $q$ in terms of a vector $p \in \mathbb{R}^{n}$, which we refer to as the potential, by analogy with the network setting, in which $p$ can be interpreted as the electric potential of the nodes. The energy of the minimum energy solution is equal to $b^{\mathrm{T}} p$. We show that the mapping $x \mapsto q$ is locally Lipschitz. Note that for $c>0$ these facts are well-known. Let us split the column indices $[m]$ of $A$ into

$$
\begin{equation*}
P \stackrel{\text { def }}{=}\left\{e \in[m]: c_{e}>0\right\} \quad \text { and } \quad Z \stackrel{\text { def }}{=}\left\{e \in[m]: c_{e}=0\right\} \tag{9.1}
\end{equation*}
$$

Lemma 9.13. Assume that every non-zero vector in the kernel of $A$ has positive cost. Let $r_{e}=c_{e} / x_{e}$ and let $R$ denote the corresponding diagonal matrix. Let us split $A$ into $A_{P}$ and $A_{Z}$, and $q$ into $q_{P}$ and $q_{Z}$. Since $A_{Z}$ has linearly independent columns, we may assume that the first $|Z|$ rows of $A_{Z}$ form a square non-singular matrix. We can thus write $A=\left[\begin{array}{cc}A_{P}^{\prime} & A_{Z}^{\prime} \\ A_{P}^{\prime \prime} & A_{Z}^{\prime \prime}\end{array}\right]$ with invertible $A_{Z}^{\prime}$. Then the minimum energy solution satisfies

$$
\left[\begin{array}{ll}
A_{P}^{\prime} & A_{Z}^{\prime}  \tag{9.2}\\
A_{P}^{\prime \prime} & A_{Z}^{\prime \prime}
\end{array}\right]\left[\begin{array}{l}
q_{P} \\
q_{Z}
\end{array}\right]=\left[\begin{array}{l}
b^{\prime} \\
b^{\prime \prime}
\end{array}\right] \quad \text { and } \quad\left[\begin{array}{cc}
R_{P} & 0 \\
0 & 0
\end{array}\right]\left[\begin{array}{l}
q_{P} \\
q_{Z}
\end{array}\right]=\left[\begin{array}{cc}
A_{P}^{\prime \mathrm{T}} & A_{P}^{\prime \prime \mathrm{T}} \\
A_{Z}^{\prime \mathrm{T}} & A_{Z}^{\prime \prime \mathrm{T}}
\end{array}\right]\left[\begin{array}{c}
p^{\prime} \\
p^{\prime \prime}
\end{array}\right]
$$

for some vector $p=\left[\begin{array}{c}p^{\prime} \\ p^{\prime \prime}\end{array}\right]$; here $p^{\prime}$ has dimension $|Z|$. The equation system (9.2) has a unique solution given by

$$
\left[\begin{array}{l}
q_{Z}  \tag{9.3}\\
q_{P}
\end{array}\right]=\left[\begin{array}{c}
{\left[A_{Z}^{\prime}\right]^{-1}\left(b^{\prime}-A_{P}^{\prime} q_{P}\right)} \\
R_{P}^{-1} A_{P}^{\mathrm{T}} p
\end{array}\right] \quad \text { and } \quad\left[\begin{array}{c}
p^{\prime} \\
p^{\prime \prime}
\end{array}\right]=\left[\begin{array}{c}
-\left[\left[A_{Z}^{\prime}\right]^{\mathrm{T}}\right]^{-1}\left[A_{Z}^{\prime \prime}\right]^{\mathrm{T}} p^{\prime \prime} \\
M R^{-1} M^{\mathrm{T}}\left(b^{\prime \prime}-A_{Z}^{\prime \prime}\left[A_{Z}^{\prime}\right]^{-1} b^{\prime}\right)
\end{array}\right],
$$

where $M=A_{P}^{\prime \prime}-A_{Z}^{\prime \prime}\left[A_{Z}^{\prime}\right]^{-1} A_{P}^{\prime}$ is the Schur complement of the block $A_{Z}^{\prime}$ of the matrix $A$.
Proof. $q$ minimizes $E(f)=f^{\mathrm{T}} R f$ among all solutions of $A f=b$. The KKT conditions state that $q$ must satisfy $R q=A^{\mathrm{T}} p$ for some $p$. Note that $2 R f$ is the gradient of the energy $E(f)$ with respect to $f$ and that the $-A^{\mathrm{T}} p$ is the gradient of $p^{\mathrm{T}}(b-A f)$ with respect to $f$. We may absorb the factor -2 in $p$. Thus $q$ satisfies (9.2).

We show next that the linear system (9.2) has a unique solution. The top $|Z|$ rows of the left system in (9.2) give

$$
\begin{equation*}
q_{Z}=\left[A_{Z}^{\prime}\right]^{-1}\left(b^{\prime}-A_{P}^{\prime} q_{P}\right) \tag{9.4}
\end{equation*}
$$

Substituting this expression for $q_{Z}$ into the bottom $n-|Z|$ rows of the left system in (9.2) yields

$$
M q_{P}=b^{\prime \prime}-A_{Z}^{\prime \prime}\left[A_{Z}^{\prime}\right]^{-1} b^{\prime}
$$

From the top $|P|$ rows of the right system in (9.2) we infer $q_{P}=R_{P}^{-1} A_{P}^{\mathrm{T}} \cdot p$. Thus

$$
\begin{equation*}
M R_{P}^{-1} A_{P}^{\mathrm{T}} \cdot p=b^{\prime \prime}-A_{Z}^{\prime \prime}\left[A_{Z}^{\prime}\right]^{-1} \cdot b^{\prime} \tag{9.5}
\end{equation*}
$$

The bottom $n-|Z|$ rows of the right system in (9.2) yield $0=A_{Z}^{\mathrm{T}} p=\left[A_{Z}^{\prime}\right]^{\mathrm{T}} p^{\prime}+\left[A_{Z}^{\prime \prime}\right]^{\mathrm{T}} p^{\prime \prime}$ and hence

$$
\begin{equation*}
p^{\prime}=-\left[\left[A_{Z}^{\prime}\right]^{\mathrm{T}}\right]^{-1}\left[A_{Z}^{\prime \prime}\right]^{\mathrm{T}} p^{\prime \prime} \tag{9.6}
\end{equation*}
$$

Substituting (9.6) into (9.5) yields

$$
\begin{align*}
b^{\prime \prime}-A_{Z}^{\prime \prime}\left[A_{Z}^{\prime}\right]^{-1} b^{\prime} & =M R_{P}^{-1}\left(\left[A_{P}^{\prime}\right]^{\mathrm{T}} p^{\prime}+\left[A_{P}^{\prime \prime}\right]^{\mathrm{T}} p^{\prime \prime}\right) \\
& =M R_{P}^{-1}\left(\left[A_{P}^{\prime \prime}\right]^{\mathrm{T}}-\left[A_{P}^{\prime}\right]^{\mathrm{T}}\left[\left[A_{Z}^{\prime}\right]^{\mathrm{T}}\right]^{-1}\left[A_{Z}^{\prime \prime}\right]^{\mathrm{T}}\right) p^{\prime \prime} \\
& =M R_{P}^{-1} M^{\mathrm{T}} p^{\prime \prime} \tag{9.7}
\end{align*}
$$

It remains to show that the matrix $M R_{P}^{-1} M^{\mathrm{T}}$ is non-singular. We first observe that the rows of $M$ are linearly independent. Consider the left system in (9.2). Multiplying the first $|Z|$ rows by $\left[A_{Z}^{\prime}\right]^{-1}$ and then subtracting $A_{Z}^{\prime \prime}$ times the resulting rows from the last $n-|Z|$ rows turns $A$ into the matrix $Q=\left[\begin{array}{ccc}{\left[A_{Z}^{\prime}\right]^{-1} A_{P}^{\prime}} & I \\ 0\end{array}\right]$. By assumption, $A$ has independent rows. Moreover, the preceding operations guarantee that $\operatorname{rank}(A)=\operatorname{rank}(Q)$. Therefore, $M$ has independent rows. Since $R_{P}^{-1}$ is a positive diagonal matrix, $R_{P}^{-1 / 2}$ exists and is a positive diagonal matrix. Let $z$ be an arbitrary non-zero vector of dimension $|P|$. Then $z^{\mathrm{T}} M R_{P}^{-1} M^{\mathrm{T}} z=\left(R_{P}^{-1 / 2} M^{\mathrm{T}} z\right)^{\mathrm{T}}\left(R_{P}^{-1 / 2} M^{\mathrm{T}} z\right)>0$ and hence $M R_{P}^{-1} M^{\mathrm{T}}$ is non-singular. It is even positive semi-definite.

There is a shorter proof that the system (9.2) has a unique solution. However, the argument does not give an explicit expression for the solution. In the case of a convex objective function and affine constraints, the KKT conditions are sufficient for being a global minimum. Thus any solution to (9.2) is a global optimum. We have already shown in Lemma 9.11 that the global minimum is unique.

We next observe that the energy of $q$ can be expressed in terms of the potential.
Lemma 9.14. Let $q$ be the minimum energy feasible solution and let $f$ be any feasible solution. Then $E(q)=b^{\mathrm{T}} p=f^{\mathrm{T}} A^{\mathrm{T}} p$.
Proof. As in the proof of Lemma 9.13, we split $q$ into $q_{P}$ and $q_{Z}, R$ into $R_{P}$ and $R_{Z}$, and $A$ into $A_{P}$ and $A_{Z}$. Then

$$
\begin{aligned}
E(q) & =q_{P}^{\mathrm{T}} R_{P} q_{P} \\
& =p^{\mathrm{T}} A_{P} q_{p} \\
& =p^{\mathrm{T}}\left(b-A_{Z} q_{Z}\right) \\
& =b^{\mathrm{T}} p
\end{aligned}
$$

by the definition of $E(q)$ and since $R_{Z}=0$
by the right system in (9.2)
by the left system in (9.2)
by the right system in (9.2).
For any feasible solution $f$, we have $f^{\mathrm{T}} A^{\mathrm{T}} p=b^{\mathrm{T}} p$.

## The Mapping $x \mapsto q$ is Locally Lipschitz

We show that the mapping $x \mapsto q$ is locally Lipschitz continuous; this implies existence of a solution $x(t)$ with domain $[0, \infty)$ by standard arguments. Our analysis builds upon Cramer's rule and the CauchyBinet formula. The Cauchy-Binet formula extends Kirchhoff's spanning tree theorem which was used in [BMV12] for the analysis of the undirected shortest path problem.
Lemma 9.15 (Local Lipschitz Condition). Assume $c \geqslant 0$, no non-zero vector in the kernel of $A$ has cost zero, and that $A, b$, and $c$ are integral. Let $\alpha, \beta>0$. For any two vectors $x$ and $\widetilde{x}$ in $\mathbb{R}^{m}$ with $\alpha \leqslant x_{e}, \widetilde{x}_{e} \leqslant \beta$ for all e, define $\gamma \stackrel{\text { def }}{=} 2 m^{n}(\beta / \alpha)^{n} c_{\max }^{n} D^{2}\left\|b / \gamma_{A}\right\|_{1}$. Then $\left\|q_{e}(x)\left|-\left|q_{e}(\widetilde{x})\right|\|\gamma\| x-\widetilde{x} \|_{\infty}\right.\right.$ for every $e \in[m]$.
Proof. First assume that $c>0$. By Cramer's rule

$$
\left(A R^{-1} A^{\mathrm{T}}\right)^{-1}=\frac{1}{\operatorname{det}\left(A R^{-1} A^{\mathrm{T}}\right)}\left((-1)^{i+j} \operatorname{det}\left(M_{-j,-i}\right)\right)_{i j},
$$

where $M_{-i,-j}$ is obtained from $A R^{-1} A^{\mathrm{T}}$ by deleting the $i$-th row and the $j$-th column. For a subset $S$ of $[m]$ and an index $i \in[n]$, let $A_{S}$ be the $n \times|S|$ matrix consisting of the columns selected by $S$ and let $A_{-i, S}$ be the matrix obtained from $A_{S}$ by deleting row $i$. If $D$ is a diagonal matrix of size $m$, then $(A D)_{S}=A_{S} D_{S}$. The Cauchy-Binet theorem expresses the determinant of a product of two matrices (not necessarily square) as a sum of determinants of square matrices. It yields

$$
\begin{aligned}
\operatorname{det}\left(A R^{-1} A^{\mathrm{T}}\right) & =\sum_{S \subseteq[m] ;|S|=n}\left(\operatorname{det}\left(\left(A R^{-1 / 2}\right)_{S}\right)\right)^{2} \\
& =\sum_{S \subseteq[m] ;|S|=n}\left(\prod_{e \in S} x_{e} / c_{e}\right) \cdot\left(\operatorname{det} A_{S}\right)^{2} .
\end{aligned}
$$

Similarly,

$$
\operatorname{det}\left(A R^{-1} A^{\mathrm{T}}\right)_{-i,-j}=\sum_{S \subseteq[m] ;|S|=n-1}\left(\prod_{e \in S} x_{e} / c_{e}\right) \cdot\left(\operatorname{det} A_{-i, S} \cdot \operatorname{det} A_{-j, S}\right)
$$

Using $p=\left(A R^{-1} A^{\mathrm{T}}\right)^{-1} b$, we obtain

$$
\begin{equation*}
p_{i}=\frac{\sum_{j \in[n]}(-1)^{i+j} \sum_{S \subseteq[m] ;|S|=n-1}\left(\prod_{e \in S} x_{e} / c_{e}\right) \cdot\left(\operatorname{det} A_{-i, S} \cdot \operatorname{det} A_{-j, S}\right) b_{j}}{\sum_{S \subseteq[m] ;|S|=n}\left(\prod_{e \in S} x_{e} / c_{e}\right) \cdot\left(\operatorname{det} A_{S}\right)^{2}} . \tag{9.8}
\end{equation*}
$$

Substituting into $q=R^{-1} A^{\mathrm{T}} p$ yields

$$
\begin{align*}
q_{e} & =\frac{x_{e}}{c_{e}} A_{e}^{\mathrm{T}} p \\
& =\frac{x_{e}}{c_{e}} \sum_{i} A_{i, e} \cdot \frac{\sum_{j \in[n]}(-1)^{i+j+2 n} \sum_{S \subseteq[m] ;|S|=n-1}\left(\prod_{e^{\prime} \in S} x_{e^{\prime}} / c_{e^{\prime}}\right) \cdot\left(\operatorname{det} A_{-i, S} \cdot \operatorname{det} A_{-j, S}\right) b_{j}}{\sum_{S \subseteq[m] ;|S|=n}\left(\prod_{e^{\prime} \in S} x_{e^{\prime}} / c_{e^{\prime}}\right) \cdot\left(\operatorname{det} A_{S}\right)^{2}} \\
& =\frac{\sum_{S \subseteq[m] ;|S|=n-1}\left(\prod_{e^{\prime} \in S \cup e} x_{e^{\prime}} / c_{e^{\prime}}\right) \cdot \sum_{i \in[n]}(-1)^{i+n} A_{i, e} \operatorname{det} A_{-i, S} \cdot \sum_{j \in[n]}(-1)^{j+n} b_{j} \operatorname{det} A_{-j, S}}{\sum_{S \subseteq[m] ;|S|=n}\left(\prod_{e^{\prime} \in S} x_{e^{\prime}} / c_{e^{\prime}}\right) \cdot\left(\operatorname{det} A_{S}\right)^{2}} \\
& =\frac{\sum_{S \subseteq[m] ;|S|=n-1}\left(\prod_{e^{\prime} \in S \cup e} x_{e^{\prime}} / c_{e^{\prime}}\right) \cdot \operatorname{det}\left(A_{S} \mid A_{e}\right) \cdot \operatorname{det}\left(A_{S} \mid b\right)}{\sum_{S \subseteq[m] ;|S|=n}\left(\prod_{e^{\prime} \in S} x_{e^{\prime}} / c_{e^{\prime}}\right) \cdot\left(\operatorname{det} A_{S}\right)^{2}}, \tag{9.9}
\end{align*}
$$

where $\left(A_{S} \mid A_{e}\right)$, respectively $\left(A_{S} \mid b\right)$, denotes the $n \times n$ matrix whose columns are selected from $A$ by $S$ and whose last column is equal to $A_{e}$, respectively $b$.

We are now ready to estimate the derivative $\partial q_{e} / \partial x_{i}$. Assume first that $e \neq i$. By the above, $q_{e}=$ $\frac{x_{e}}{c_{e}} \frac{F+G x_{i} / c_{i}}{H+I x_{i} / c_{i}}$, where $F, G, H$ and $I$ are given implicitly by (9.9). Then

$$
\left|\frac{\partial q_{e}}{\partial x_{i}}\right|=\left|\frac{x_{e}}{c_{e}} \cdot \frac{F I / c_{i}-G H / c_{i}}{\left(H+I x_{i} / c_{i}\right)^{2}}\right| \leqslant \frac{2 \cdot\binom{m}{n-1} \beta^{n} D^{2}\left\|b / \gamma_{A}\right\|_{1}}{\left(\alpha / c_{\max }\right)^{n}} \leqslant \gamma
$$

For $e=i$, we have $q_{e}=\frac{G x_{e} / c_{e}}{H+I x_{e} / c_{e}}$, where $G, H$, and $I$ are given implicitly by (9.9). Then

$$
\left|\frac{\partial q_{e}}{\partial x_{e}}\right|=\left|\frac{G H / c_{e}}{\left(H+I x_{e} / c_{e}\right)^{2}}\right| \leqslant \frac{\binom{m}{n-1} \beta^{n} D^{2}\left\|b / \gamma_{A}\right\|_{1}}{\left(\alpha / c_{\max }\right)^{n}} \leqslant \gamma
$$

Finally, consider $x$ and $\widetilde{x}$ with $\alpha \leqslant x_{e}, \widetilde{x}_{e} \leqslant \beta$ for all $e$. Let $\bar{x}_{\ell}=\left(\widetilde{x}_{1}, \ldots, \widetilde{x}_{\ell}, x_{\ell+1}, \ldots, x_{m}\right)$. Then

$$
\left\|q _ { e } ( x ) \left|-\left|q_{e}(\widetilde{x})\left\|\leqslant\left|q_{e}(x)-q_{e}(\widetilde{x})\right| \leqslant \sum_{0 \leqslant \ell<m}\left|q_{e}\left(\bar{x}_{\ell}\right)-q_{e}\left(\bar{x}_{\ell+1}\right)\right| \leqslant \gamma\right\| x-\widetilde{x} \|_{1}\right.\right.\right.
$$

In the general case where $c \geqslant 0$, we first derive an expression for $p^{\prime \prime}$ similar to (9.8). Then the equations for $p^{\prime}$ in (9.3) yield $p^{\prime}$, the equations for $q_{P}$ in (9.3) yield $q_{P}$, and finally the equations for $q_{Z}$ in (9.3) yield $q_{Z}$.

We are now ready to establish the existence of a solution with domain $[0, \infty)$.
Lemma 9.16. The solution to the undirected dynamics in (8.1) has domain $[0, \infty)$. Moreover, for every $t \geqslant 0$ and $e \in[m]$, we have

$$
x_{e}(0) \cdot \exp \{-t\} \leqslant x_{e}(t) \leqslant D\left\|b / \gamma_{A}\right\|_{1}+\max \left(0, x_{e}(0)-D\left\|b / \gamma_{A}\right\|_{1}\right) \cdot \exp \{-t\}
$$

Proof. Consider any $x_{0}>0$ and any $t_{0} \geqslant 0$. We first show that there is a positive $\delta^{\prime}$ (depending on $x_{0}$ ) such that a unique solution $x(t)$ with $x\left(t_{0}\right)=x_{0}$ exists for $t \in\left(t_{0}-\delta^{\prime}, t_{0}+\delta^{\prime}\right)$. By the Picard-Lindelöf Theorem [Tes12, Theorem 2.2], this holds true if the mapping $x \mapsto|q|-x$ is continuous and satisfies a Lipschitz condition in a neighborhood of $x_{0}$. Continuity clearly holds. Let $\varepsilon=\min _{i}\left(x_{0}\right)_{i} / 2$ and let $U=\left\{x:\left\|x-x_{0}\right\|_{\infty}<\varepsilon\right\}$. Then for every $x, \widetilde{x} \in U$ and every $e$

$$
\left|\left|q_{e}(x)\right|-\left|q_{e}(\widetilde{x})\right|\right| \leqslant \gamma\|x-\widetilde{x}\|_{1}
$$

where $\gamma$ is as in Lemma 9.15. Local existence implies the existence of a solution which cannot be extended. Since $q$ is bounded (Lemma 9.12), $x$ is bounded at all finite times, and hence the solution exists for all $t$. The lower bound $x_{e}(t) \geqslant x_{e}(0) \cdot \exp \{-t\}>0$ for all $e$, holds by Lemma 9.2 with $A=0$ and $\alpha=-1$. Since $\left|q_{e}\right| \leqslant$ $D\left\|b / \gamma_{A}\right\|_{1}, \dot{x}_{e}=\left|q_{e}\right|-x_{e} \leqslant D\left\|b / \gamma_{A}\right\|_{1}-x_{e}$, we have $x_{e}(t) \leqslant D\left\|b / \gamma_{A}\right\|_{1}+\max \left(0, x_{e}(0)-D\left\|b / \gamma_{A}\right\|_{1}\right) \cdot \exp \{-t\}$ by Lemma 9.2 with $B=D\left\|b / \gamma_{A}\right\|_{1}$ and $\beta=-1$.

### 9.2.2 LP Duality

The energy $E(x)$ is no longer a Lyapunov function, e.g., if $x(0) \approx \mathbf{0}, x(t)$ and hence $E(x(t))$ will grow initially. We will show that energy suitably scaled is a Lyapunov function. What is the appropriate scaling factor? In the case of the undirected shortest path problem, [BMV12] used the minimum capacity of any source-sink cut as a scaling factor. The proper generalization to our situation is to consider the linear program $\max \{\alpha: A f=\alpha b,|f| \leqslant x\}$, where $x$ is a fixed positive vector. Linear programming duality yields the corresponding minimization problem which generalizes the minimum cut problem to our situation.

Lemma 9.17. Let $x \in \mathbb{R}_{>0}^{m}$ and $b \neq 0$. The linear programs

$$
\begin{equation*}
\max \{\alpha: A f=\alpha b,|f| \leqslant x\} \quad \text { and } \quad \min \left\{\left|y^{\mathrm{T}} A\right| x: b^{\mathrm{T}} y=-1\right\} \tag{9.10}
\end{equation*}
$$

are feasible and have the same objective value. Moreover, there is a finite set $Y_{A}=\left\{d^{1}, \ldots, d^{K}\right\}$ of vectors $d^{i} \in \mathbb{R}_{\geqslant 0}^{m}$ that are independent of $x$ such that the minimum above is equal to $C_{\star}=\min _{d \in Y_{A}} d^{\mathrm{T}} x$. There is a feasible $f$ with $|f| \leqslant x / C_{\star} .{ }^{2}$

[^14]Proof. The pair $(\alpha, f)=(0,0)$ is a feasible solution for the maximization problem. Since $b \neq 0$, there exists $y$ with $b^{\mathrm{T}} y=-1$ and thus both problems are feasible. The dual of $\max \{\alpha: A f-\alpha b=0, f \leqslant x,-f \leqslant x\}$ has unconstrained variables $y \in \mathbb{R}^{n}$ and non-negative variables $z^{+}, z^{-} \in \mathbb{R}^{m}$ and reads

$$
\begin{equation*}
\min \left\{x^{\mathrm{T}}\left(z^{+}+z^{-}\right):-b^{\mathrm{T}} y=1, A^{\mathrm{T}} y+z^{+}-z^{-}=0, z^{+}, z^{-} \geqslant 0\right\} . \tag{9.11}
\end{equation*}
$$

From $z^{-}=A^{\mathrm{T}} y+z^{+}, z^{+} \geqslant 0, z^{-} \geqslant 0$ and $x>0$, we conclude $\min \left(z^{+}, z^{-}\right)=0$ in an optimal solution. Thus $z^{-}=\max \left(0, A^{\mathrm{T}} y\right)$ and $z^{+}=\max \left(0,-A^{\mathrm{T}} y\right)$ and hence $z^{+}+z^{-}=\left|A^{\mathrm{T}} y\right|$ in an optimal dual solution. Therefore, (9.11) and the right LP in (9.10) have the same objective value.

We next show that the dual attains its minimum at a vertex of the feasible set. For this it suffices to show that its feasible set contains no line. Assume it does. Then there are vectors $d=\left(y_{1}, z_{1}^{+}, z_{1}^{-}\right), d$ non-zero, and $p=\left(y_{0}, z_{0}^{+}, z_{0}^{-}\right)$such that $\left(y, z^{+}, z^{-}\right)=p+\lambda d=\left(y_{0}+\lambda y_{1}, z_{0}^{+}+\lambda z_{1}^{+}, z_{0}^{-}+\lambda z_{1}^{-}\right)$is feasible for all $\lambda \in \mathbb{R}$. Thus $z_{1}^{+}=z_{1}^{-}=0$. Note that if either $z_{1}^{+}$or $z_{1}^{-}$were non-zero then either $z_{0}^{+}+\lambda z_{1}^{+}$or $z_{0}^{-}+\lambda z_{1}^{-}$would have a negative component for some $\lambda$. Then $A^{\mathrm{T}} y+z^{+}+z^{-}=0$ implies $A^{\mathrm{T}} y_{1}=0$. Since $A$ has full row rank, $y_{1}=0$. Thus the dual contains no line and the minimum is attained at a vertex of its feasible region. The feasible region of the dual does not depend on $x$.

Let $\left(y^{1}, z_{1}^{+}, z_{1}^{-}\right)$to $\left(y^{K}, z_{K}^{+}, z_{K}^{-}\right)$be the vertices of (9.11), and let $Y_{A}=\left\{\left|A^{\mathrm{T}} y^{1}\right|, \ldots,\left|A^{\mathrm{T}} y^{K}\right|\right\}$. Then

$$
\begin{aligned}
\min _{d \in Y_{A}} d^{\mathrm{T}} x & =\min \left\{x^{\mathrm{T}}\left(z^{+}+z^{-}\right):-b^{\mathrm{T}} y=1, A^{\mathrm{T}} y+z^{+}-z^{-}=0, z^{+}, z^{-} \geqslant 0\right\} \\
& =\min \left\{\left|y^{\mathrm{T}} A\right| x: b^{\mathrm{T}} y=-1\right\} .
\end{aligned}
$$

We finally show that there is a feasible $f$ with $|f| \leqslant x / C_{\star}$. Let $x^{\prime} \stackrel{\text { def }}{=} x / C_{\star}$. Then $x^{\prime}>0$ and $\min _{d \in Y_{A}} d^{\mathrm{T}} x^{\prime}=\min _{d \in Y_{A}} d^{\mathrm{T}} x / C_{\star}=C_{\star} / C_{\star}=1$ and thus the right LP with $x=x^{\prime}$ (9.10) has objective value 1 . Hence, the left LP has objective value 1 and there is a feasible $f$ with $|f| \leqslant x^{\prime}$.

### 9.2.3 Convergence to Dominance

In the network setting, an important role is played by the set of edge capacity vectors that support a feasible flow. In the LP setting, we generalize this notion to the set of dominating states, which is defined as

$$
X_{\text {dom }} \stackrel{\text { def }}{=}\left\{x \in \mathbb{R}^{m}: \exists \text { feasible } f:|f| \leqslant x\right\}
$$

An alternative characterization, using the set $Y_{A}$ from Lemma 9.17, is

$$
\mathcal{X}_{1} \stackrel{\text { def }}{=}\left\{x \in \mathbb{R}_{\geqslant 0}^{m}: d^{\mathrm{T}} x \geqslant 1 \text { for all } d \in Y_{A}\right\} .
$$

We now prove that $X_{\text {dom }}=\mathcal{X}_{1}$ and that the set $\mathcal{X}_{1}$ is an attractor in the following sense.
Lemma 9.18. The following statements hold:
(i) $X_{\text {dom }}=\mathcal{X}_{1}$. Moreover, $\lim _{t \rightarrow \infty} \operatorname{dist}\left(x(t), \mathcal{X}_{1}\right)=0$, where $\operatorname{dist}\left(x, \mathcal{X}_{1}\right)$ is the Euclidean distance between $x$ and $\mathcal{X}_{1}$.
(ii) If $x\left(t_{0}\right) \in \mathcal{X}_{1}$, then $x(t) \in \mathcal{X}_{1}$ for all $t \geqslant t_{0}$. For all sufficiently large $t, x(t) \in \mathcal{X}_{1 / 2} \stackrel{\text { def }}{=}\left\{x \in \mathbb{R}_{\geqslant 0}^{n}\right.$ : $d^{\mathrm{T}} x \geqslant 1 / 2$ for all $\left.d \in Y_{A}\right\}$, and if $x \in \mathcal{X}_{1 / 2}$ then there is a feasible $f$ with $|f| \leqslant 2 x$.
Proof. (i) If $x \in \mathcal{X}_{1}$, then $d^{\mathrm{T}} x \geqslant 1$ for all $d \in Y_{A}$ and hence Lemma 9.17 implies the existence of a feasible solution $f$ with $|f| \leqslant x$. Conversely, if $x \in X_{\text {dom }}$, then there is a feasible $f$ with $|f| \leqslant x$. Thus $d^{\mathrm{T}} x \geqslant 1$ for all $d \in Y_{A}$ and hence $x \in \mathcal{X}_{1}$. By the proof of Lemma 9.17, for any $d \in Y_{A}$, there is a $y$ such that $d=\left|A^{\mathrm{T}} y\right|$ and $b^{\mathrm{T}} y=-1$. Let $Y(t)=d^{\mathrm{T}} x$. Then

$$
\dot{Y}=\left|y^{\mathrm{T}} A\right| \dot{x}=\left|y^{\mathrm{T}} A\right|(|q|-x) \geqslant\left|y^{\mathrm{T}} A q\right|-Y=\left|y^{\mathrm{T}} b\right|-Y=1-Y
$$

Thus for any $t_{0}$ and $t \geqslant t_{0}, Y(t) \geqslant 1+\left(Y\left(t_{0}\right)-1\right) e^{-\left(t-t_{0}\right)}$ by Lemma 9.2 applied with $A=1$ and $\alpha=-1$. In particular, $\liminf _{t \rightarrow \infty} Y(t) \geqslant 1$. Thus $\liminf _{t \rightarrow \infty} \min _{d \in Y_{A}} d^{\mathrm{T}} x \geqslant 1$ and hence $\lim _{t \rightarrow \infty} \operatorname{dist}\left(x(t), \mathcal{X}_{1}\right)=0$.
(ii) Moreover, if $Y\left(t_{0}\right) \geqslant 1$, then $Y(t) \geqslant 1$ for all $t \geqslant t_{0}$. Hence $x\left(t_{0}\right) \in \mathcal{X}_{1}$ implies $x(t) \in \mathcal{X}_{1}$ for all $t \geqslant t_{0}$. Since $x(t)$ converges to $\mathcal{X}_{1}, x(t) \in \mathcal{X}_{1 / 2}$ for all sufficiently large $t$. If $x \in \mathcal{X}_{1 / 2}$ there is $f$ such that $A f=\frac{1}{2} b$ and $|f| \leqslant x$. Thus $2 f$ is feasible and $|2 f| \leqslant 2 x$.

The next lemma summarizes simple bounds on the values of resistors $r$, potentials $p$ and states $x$ that hold for sufficiently large $t$. Recall that $P=\left\{e \in[m]: c_{e}>0\right\}$ and $Z=\left\{e \in[m]: c_{e}=0\right\}$, see (9.1).

Lemma 9.19. The following statements hold:
(i) For sufficiently large $t$, it holds that $r_{e} \geqslant c_{e} /\left(2 D\left\|b / \gamma_{A}\right\|_{1}\right)$, $b^{\mathrm{T}} p \leqslant 8 D\left\|b / \gamma_{A}\right\|_{1}\|c\|_{1}$ and $\left|A_{e}^{\mathrm{T}} p\right| \leqslant$ $8 D^{2}\|b\|_{1}\|c\|_{1}$ for all $e$.
(ii) For all $e$, it holds that $\dot{x}_{e} / x_{e} \geqslant-1$ and for all $e \in P$, it holds that $\dot{x}_{e} / x_{e} \leqslant 8 D^{2}\|b\|_{1}\|c\|_{1} / c_{\text {min }}$.
(iii) There is a positive constant $C$ such that for all $t \geqslant t_{0}$, there is a feasible $f$ (depending on $t$ ) such that $x_{e}(t) \geqslant C$ for all indices $e$ in the support of $f$.

Proof. (i) By Lemma 9.16, $x_{e}(t) \leqslant 2 D\left\|b / \gamma_{A}\right\|_{1}$ for all sufficiently large $t$. It follows that $r_{e}=c_{e} / x_{e} \geqslant$ $c_{e} /\left(2 D\left\|b / \gamma_{A}\right\|_{1}\right)$. Due to Lemma 9.18, for large enough $t$, there is a feasible flow with $|f| \leqslant 2 x$. Together with $x_{e}(t) \leqslant 2 D\left\|b / \gamma_{A}\right\|_{1}$, it follows that

$$
b^{\mathrm{T}} p=E(q) \leqslant E(2 x)=4 c^{\mathrm{T}} x \leqslant 8 D\left\|b / \gamma_{A}\right\|_{1}\|c\|_{1}
$$

Now, orient $A$ according to $q$ and consider any index $e^{\prime}$. Recall that for all indices $e$, we have $A_{e}^{\mathrm{T}} p=0$ if $e \in Z$, and $q_{e}=\left(x_{e} / c_{e}\right) \cdot A_{e}^{\mathrm{T}} p$ if $e \in P$. Thus $A_{e}^{\mathrm{T}} p \geqslant 0$ for all $e$. If $e^{\prime} \in Z$ or $e^{\prime} \in P$ and $q_{e^{\prime}}=0$, the claim is obvious. So assume $e^{\prime} \in P$ and $q_{e^{\prime}}>0$. Since $q$ is a convex combination of $q$-sign-compatible basic feasible solutions, there is a basic feasible solution $f$ with $f \geqslant 0$ and $f_{e^{\prime}}>0$. By Lemma 9.10, $f_{e^{\prime}} \geqslant 1 /\left(D \gamma_{A}\right)$. Therefore

$$
f_{e^{\prime}} A_{e^{\prime}}^{\mathrm{T}} p \leqslant \sum_{e} f_{e} A_{e}^{\mathrm{T}} p=b^{\mathrm{T}} p \leqslant 8 D\left\|b / \gamma_{A}\right\|_{1}\|c\|_{1}
$$

for all sufficiently large $t$. The inequality follows from $f_{e} \geqslant 0$ and $A_{e}^{\mathrm{T}} p \geqslant 0$ for all $e$. Thus $A_{e^{\prime}}^{\mathrm{T}} p \leqslant$ $8 D^{2}\|b\|_{1}\|c\|_{1}$ for all sufficiently large $t$.
(ii) We have $\dot{x}_{e} / x_{e}=\left(\left|q_{e}\right|-x_{e}\right) / x_{e} \geqslant-1$ for all $e$. For $e$ with $c_{e}>0$

$$
\frac{\dot{x}_{e}}{x_{e}}=\frac{\left|q_{e}\right|-x_{e}}{x_{e}} \leqslant \frac{\left|A_{e}^{\mathrm{T}} p\right|}{c_{e}} \leqslant 8 D^{2}\|b\|_{1}\|c\|_{1} / c_{\mathrm{min}}
$$

(iii) Let $t_{0}$ be such that $d^{\mathrm{T}} x(t) \geqslant 1 / 2$ for all $d \in Y_{A}$ and $t \geqslant t_{0}$. Then for all $t \geqslant t_{0}$, there is $f$ such that $A f=\frac{1}{2} b$ and $|f| \leqslant x(t) ; f$ may depend on $t$. By Lemma 9.1, we can write $2 f$ as convex combination of $f$-sign-compatible basic feasible solutions (at most $m$ of them) and a $f$-sign-compatible solution in the kernel of $A$. Dropping the solution in the kernel of $A$ leaves us with a solution which is still dominated by $x$.
It holds that for every $e \in E$ with $f_{e} \neq 0$, there is a basic feasible solution $g$ used in the convex decomposition such that $2\left|f_{e}\right| \geqslant\left|g_{e}\right|>0$. By Lemma 9.10, every non-zero component of $g$ is at least $1 /\left(D \gamma_{A}\right)$. We conclude that $x_{e} \geqslant 1 /\left(2 D \gamma_{A}\right)$, for every $e$ in the support of $g$.

### 9.2.4 The Equilibrium Points

We next characterize the equilibrium points

$$
\begin{equation*}
F=\left\{x \in \mathbb{R}_{\geqslant 0}:|q|=x\right\} \tag{9.12}
\end{equation*}
$$

Let us first elaborate on the special case of the undirected shortest path problem. Here the equilibria are the flows of value one from source to sink in a network formed by undirected source-sink paths of the same length. This can be seen as follows. Consider any $x \geqslant 0$ and assume $\operatorname{supp}(x)$ is a network of undirected source-sink paths of the same length. Call this network $\mathcal{N}$. Assign to each node $u$, a potential $p_{u}$ equal to the length of the shortest undirected path from the sink $s_{1}$ to $u$. These potentials are welldefined as all paths from $s_{1}$ to $u$ in $\mathcal{N}$ must have the same length. For an edge $e=(u, v)$ in $\mathcal{N}$, we have $q_{e}=x_{e} / c_{e}\left(p_{u}-p_{v}\right)=x_{e} / c_{e} \cdot c_{e}=x_{e}$, i.e., $q=x$ is the electrical flow with respect to the resistances $c_{e} / x_{e}$. Conversely, if $x$ is an equilibrium point and the network is oriented such that $q \geqslant 0$, we have $x_{e}=q_{e}=x_{e} / c_{e}\left(p_{u}-p_{v}\right)$ for all edges $e=(u, v) \in \operatorname{supp}(x)$. Thus $c_{e}=p_{u}-p_{v}$ and this is only possible if for every node $u$, all paths from $u$ to the sink have the same length. Thus $\operatorname{supp}(x)$ must be a network of undirected source-sink paths of the same length. We next generalize this reasoning.

Theorem 9.20. If $x=|q|$ is an equilibrium point and the columns of $A$ are oriented such that $q \geqslant 0$, then all feasible solutions $f$ with $\operatorname{supp}(f) \subseteq \operatorname{supp}(x)$ satisfy $c^{\mathrm{T}} f=c^{\mathrm{T}} x$. Conversely, if $x=|q|$ for a feasible $q$, $A$ is oriented such that $q \geqslant 0$, and all feasible solutions $f$ with $\operatorname{supp}(f) \subseteq \operatorname{supp}(x)$ satisfy $c^{\mathrm{T}} f=c^{\mathrm{T}} x$, then $x$ is an equilibrium point.

Proof. If $x$ is an equilibrium point, $\left|q_{e}\right|=x_{e}$ for every $e$. By changing the signs of some columns of $A$, we may assume $q \geqslant 0$, i.e., $q=x$. Let $p$ be the potential with respect to $x$. For every index $e \in P$ in the support of $x$, since $c_{e}>0$ we have $q_{e}=\frac{x_{e}}{c_{e}} A_{e}^{\mathrm{T}} p$ and hence $c_{e}=A_{e}^{\mathrm{T}} p$. Further, for the indices $e \in Z$ in the support of $x$, we have $c_{e}=0=A_{e}^{\mathrm{T}} p$ due to the second block of equations on the right hand side in (9.2). Let $f$ be any feasible solution whose support is contained in the support of $x$. Then the first part follows by

$$
\sum_{e \in \operatorname{supp}(f)} c_{e} f_{e}=\sum_{e \in \operatorname{supp}(f)} f_{e} A_{e}^{\mathrm{T}} p=b^{\mathrm{T}} p=E(q)=E(x)=\operatorname{cost}(x)
$$

For the second part, we misuse notation and use $A$ to also denote the submatrix of the constraint matrix indexed by the columns in the support of $x$. We may assume that the rows of $A$ are independent. Otherwise, we simply drop redundant constraints. We may assume $q \geqslant 0$; otherwise we simply change the sign of some columns of $A$. Then $x$ is feasible. Let $A_{B}$ be a square non-singular submatrix of $A$ and let $A_{N}$ consist of the remaining columns of $A$. The feasible solutions $f$ with $\operatorname{supp}(f) \subseteq \operatorname{supp}(x)$ satisfy $A_{B} f_{B}+A_{N} f_{N}=b$ and hence $f_{B}=A_{B}^{-1}\left(b-A_{N} f_{N}\right)$. Then

$$
c^{\mathrm{T}} f=c_{B}^{\mathrm{T}} f_{B}+c_{N}^{\mathrm{T}} f_{N}=c_{B} A_{B}^{-1} b+\left(c_{N}^{\mathrm{T}}-c_{B}^{\mathrm{T}} A_{B}^{-1} A_{N}\right) f_{N}
$$

Since, by assumption, $c^{\mathrm{T}} f$ is constant for all feasible solutions whose support is contained in the support of $x$, we must have $c_{N}=A_{N}^{\mathrm{T}}\left[A_{B}^{-1}\right]^{\mathrm{T}} c_{B}$. Let $p=\left[A_{B}^{-1}\right]^{\mathrm{T}} c_{B}$. Then it follows that $A^{\mathrm{T}} p=\left[\begin{array}{c}A_{B}^{\mathrm{T}} \\ A_{N}^{\mathrm{T}}\end{array}\right]\left[A_{B}^{-1}\right]^{\mathrm{T}} c_{B}=$ $\left[\begin{array}{c}c_{B} \\ c_{N}\end{array}\right]$ and hence $R x=A^{\mathrm{T}} p$. Thus the pair $(x, p)$ satisfies the right hand side of (9.2). Since $x$ is feasible, it also satisfies the left hand side of (9.2). Therefore, $x$ is the minimum energy solution with respect to $x$.

Corollary 9.21. Let $g$ be a basic feasible solution. Then $|g|$ is an equilibrium point.
Proof. Let $g$ be a basic feasible solution. Orient $A$ such that $g \geqslant 0$. Since $g$ is basic, there is a $B \subseteq[m]$ such that $g=\left(g_{B}, g_{N}\right)=\left(A_{B}^{-1} b, 0\right)$. Consider any feasible solution $f$ with $\operatorname{supp}(f) \subseteq \operatorname{supp}(g)$. Then $f=\left(f_{B}, 0\right)$ and hence $b=A f=A_{B} f_{B}$. Therefore, $f_{B}=g_{B}$ and hence $c^{\mathrm{T}} f=c^{\mathrm{T}} g$. Thus $x=|g|$ is an equilibrium point.

This characterization of equilibria has an interesting consequence.
Lemma 9.22. The set $L \stackrel{\text { def }}{=}\left\{c^{\mathrm{T}} x: x \in F\right\}$ of costs of equilibria is finite.
Proof. If $x$ is an equilibrium, $x=|q|$, where $q$ is the minimum energy solution with respect to $x$. Orient $A$ such that $q \geqslant 0$. Then by Theorem $9.20, c^{\mathrm{T}} f=c^{\mathrm{T}} x$ for all feasible solutions $f$ with $\operatorname{supp}(f) \subseteq \operatorname{supp}(x)$. In particular, this holds true for all such basic feasible solutions $f$. Thus $L$ is a subset of the set of costs of all basic feasible solutions, which is a finite set.

We conclude this part by showing that the optimal solutions of the undirected linear program (8.2) are equilibria.

Theorem 9.23. Let $x$ be an optimal solution to (8.2). Then $x$ is an equilibrium.
Proof. By definition, there is a feasible $f$ with $|f|=x$. Let us reorient the columns of $A$ such that $f \geqslant 0$ and let us delete all columns $e$ of $A$ with $f_{e}=0$. Consider any feasible $g$ with $\operatorname{supp}(g) \subseteq \operatorname{supp}(x)$. We claim that $c^{\mathrm{T}} x=c^{\mathrm{T}} g$. Assume otherwise and consider the point $y=x+\lambda(g-x)$. If $|\lambda|$ is sufficiently small, $y \geqslant 0$. Furthermore, $y$ is feasible and $c^{\mathrm{T}} y=c^{\mathrm{T}} x+\lambda\left(c^{\mathrm{T}} g-c^{\mathrm{T}} x\right)$. If $c^{\mathrm{T}} g \neq c^{\mathrm{T}} x, x$ is not an optimal solution to (8.2). The claim now follows from Theorem 9.20.

### 9.2.5 Convergence

In order to show convergence, we construct a Lyapunov function. The following functions play a crucial role in our analysis. Let $C_{d}=d^{\mathrm{T}} x$ for $d \in Y_{A}$, and recall that $C_{\star}=\min _{d \in Y_{A}} d^{\mathrm{T}} x$ denotes the optimum. Moreover, we define

$$
h(t) \stackrel{\text { def }}{=} \sum_{e} r_{e}\left|q_{e}\right| \frac{x_{e}}{C_{\star}}-E\left(\frac{x}{C_{\star}}\right) \quad \text { and } \quad V_{d} \stackrel{\text { def }}{=} \frac{c^{\mathrm{T}} x}{C_{d}} \text { for every } d \in Y_{A}
$$

Theorem 9.24. (1) For every $d \in Y_{A}, \dot{C}_{d} \geqslant 1-C_{d}$. Thus, if $C_{d}<1$ then $\dot{C}_{d}>0$.
(2) If $x(t) \in \mathcal{X}_{1}$, then $\frac{d}{d t} \operatorname{cost}(x(t)) \leqslant 0$ with equality if and only if $x=|q|$.
(3) Let $d \in Y_{A}$ be such that $C_{\star}=d^{\mathrm{T}} x$ at time $t$. Then it holds that $\dot{V}_{d} \leqslant h(t)$.
(4) It holds that $h(t) \leqslant 0$ with equality if and only if $|q|=\frac{x}{C_{*}}$.

Proof. (i) Recall that for $d \in Y_{A}$, there is a $y$ such that $b^{\mathrm{T}} y=-1$ and $d=\left|A^{\mathrm{T}} y\right|$. Thus $\dot{C}_{d}=$ $d^{\mathrm{T}}(|q|-x) \geqslant\left|y^{\mathrm{T}} A q\right|-C_{d}=1-C_{d}$ and hence $\dot{C}_{d}>0$, whenever $C_{d}<1$.
(ii) Remember that $E(x)=\operatorname{cost}(x)$ and that $x(t) \in \mathcal{X}_{1}$ implies that there is a feasible $f$ with $|f|=x$. Thus $E(q) \leqslant E(f) \leqslant E(x)$. Let $R$ be the diagonal matrix of entries $c_{e} / x_{e}$. Then

$$
\begin{align*}
\frac{d}{d t} \operatorname{cost}(x) & =c^{\mathrm{T}}(|q|-x)  \tag{8.1}\\
& =x^{\mathrm{T}} R^{1 / 2} R^{1 / 2}|q|-x^{\mathrm{T}} R x \\
& \leqslant\left(q^{\mathrm{T}} R q\right)^{1 / 2}\left(x^{\mathrm{T}} R x\right)^{1 / 2}-x^{\mathrm{T}} R x \\
& \leqslant 0
\end{align*}
$$

$$
\text { since } c=R x
$$

by Cauchy-Schwarz
since $E(q) \leqslant E(x)$.
If the derivative is zero, both inequalities above have to be equalities. This is only possible if the vectors $|q|$ and $x$ are parallel and $E(q)=E(x)$. Let $\lambda$ be such that $|q|=\lambda x$. Then $E(q)=\sum_{e} \frac{c_{e}}{x_{e}} q_{e}^{2}=$ $\lambda^{2} \sum_{e} c_{e} x_{e}=\lambda^{2} E(x)$. Since $E(x)>0$, this implies $\lambda=1$.
(iii) By definition of $d, C_{\star}=C_{d}$. By the first two items, we have $\dot{C}_{\star}=d^{\mathrm{T}}|q|-C_{\star}$ and $\frac{d}{d t} \operatorname{cost}(x)=$ $c^{\mathrm{T}}|q|-\operatorname{cost}(x)$. Thus

$$
\begin{aligned}
\frac{d}{d t} \frac{\operatorname{cost}(x)}{C_{\star}} & =\frac{C_{\star} \frac{d}{d t} \operatorname{cost}(x)-\dot{C}_{\star} \operatorname{cost}(x)}{C_{\star}^{2}}=\frac{C_{\star}\left(c^{\mathrm{T}}|q|-\operatorname{cost}(x)\right)-\left(d^{\mathrm{T}}|q|-C_{\star}\right) \operatorname{cost}(x)}{C_{\star}^{2}} \\
& =\frac{C_{\star} \cdot c^{\mathrm{T}}|q|-d^{\mathrm{T}}|q| \cdot c^{\mathrm{T}} x}{C_{\star}^{2}} \leqslant \sum_{e} r_{e}\left|q_{e}\right| \frac{x_{e}}{C_{\star}}-\sum_{e} r_{e}\left(\frac{x_{e}}{C_{\star}}\right)^{2}=h(t)
\end{aligned}
$$

where we used $r_{e}=c_{e} / x_{e}$ and hence $c^{\mathrm{T}}|q|=\sum_{e} r_{e} x_{e}\left|q_{e}\right|, c^{\mathrm{T}} x=E(x)$, and $d^{\mathrm{T}}|q| \geqslant\left|y^{\mathrm{T}} A q\right|=1$ since $d=\left|y^{\mathrm{T}} A\right|$ for some $y$ with $b^{\mathrm{T}} y=-1$.
(iv) We have

$$
\sum_{e} r_{e} \frac{x_{e}}{C_{\star}}\left|q_{e}\right|=\sum_{e} r_{e}^{1 / 2} \frac{x_{e}}{C_{\star}} r_{e}^{1 / 2}\left|q_{e}\right| \leqslant\left(\sum_{e} r_{e}\left(\frac{x_{e}}{C_{\star}}\right)^{2}\right)^{1 / 2}\left(\sum_{e} r_{e} q_{e}^{2}\right)^{1 / 2}=\mathbb{E}\left(\frac{x}{C_{\star}}\right)^{1 / 2} \mathbb{E}(q)^{1 / 2}
$$

by Cauchy-Schwarz. Since $h(t)=\sum_{e} r_{e}\left|q_{e}\right| \frac{x_{e}}{C_{\star}}-E\left(\frac{x}{C_{\star}}\right)$ by definition, it follows that

$$
h(t) \leqslant \mathbb{E}\left(\frac{x}{C_{\star}}\right)^{1 / 2} \cdot \mathbb{E}(q)^{1 / 2}-\mathbb{E}\left(\frac{x}{C_{\star}}\right)=E\left(\frac{x}{C_{\star}}\right)^{1 / 2} \cdot\left(\mathbb{E}(q)^{1 / 2}-\mathbb{E}\left(\frac{x}{C_{\star}}\right)^{1 / 2}\right) \leqslant 0
$$

since $x / C_{\star}$ dominates a feasible solution and hence $\mathbb{E}(q) \leqslant \mathbb{E}\left(x / C_{\star}\right)$. If $h(t)=0$, we must have equality in the application of Cauchy-Schwarz, i.e., the vectors $x / C_{\star}$ and $|q|$ must be parallel, and we must have $E(q)=E\left(x / C_{\star}\right)$ as in the proof of part ii.

We show now convergence against the set of equilibrium points. We need the following technical Lemma from [BMV12].

Lemma 9.25 (Lemma 9 in [BMV12]). Let $f(t)=\max _{d \in Y_{A}} f_{d}(t)$, where each $f_{d}$ is continuous and differentiable. If $\dot{f}(t)$ exists, then there is a $d \in Y_{A}$ such that $f(t)=f_{d}(t)$ and $\dot{f}(t)=\dot{f}_{d}(t)$.
Theorem 9.26. All trajectories converge to the set $F$ of equilibrium points.
Proof. We distinguish cases according to whether the trajectory ever enters $\mathcal{X}_{1}$ or not. If the trajectory enters $\mathcal{X}_{1}$, say $x\left(t_{0}\right) \in \mathcal{X}_{1}$, then $\frac{d}{d t} \operatorname{cost}(x) \leqslant 0$ for all $t \geqslant t_{0}$ with equality only if $x=|q|$. Thus the trajectory converges to the set of fix points. If the trajectory never enters $\mathcal{X}_{1}$, consider $V=\max _{d \in Y_{A}}\left(V_{d}+1-C_{d}\right)$. We show that $\dot{V}$ exists for almost all $t$. Moreover, if $\dot{V}(t)$ exists, then $\dot{V}(t) \leqslant 0$ with equality if and only if $\left|q_{e}\right|=x_{e}$ for all $e$. It holds that $V$ is Lipschitz continuous as the maximum of a finite number of continuously differentiable functions. Since $V$ is Lipschitz continuous, the set of $t$ 's where $\dot{V}(t)$ does not exist has zero Lebesgue measure (see for example [CLSW98, Ch. 3]). If $\dot{V}(t)$ exists, we have $\dot{V}(t)=\dot{V}_{d}(t)-\dot{C}_{d}(t)$ for some $d \in Y_{A}$ according to Lemma 9.25. Then, it holds that $\dot{V}(t) \leqslant h(t)-\left(1-C_{d}\right) \leqslant 0$. Thus $x(t)$ converges to the set

$$
\left\{x \in \mathbb{R}_{\geqslant 0}: \dot{V}=0\right\}=\left\{x \in \mathbb{R}_{\geqslant 0}:|q|=x / C \text { and } C=1\right\}=\left\{x \in \mathbb{R}_{\geqslant 0}:|q|=x\right\} .
$$

At this point, we know that all trajectories $x(t)$ converge to $F$. Our next goal is to show that $c^{\mathrm{T}} x(t)$ converges to the cost of an optimum solution of (8.2) and that $|q|-x$ converges to zero. We are only able to show the latter for all indices $e \in P$, i.e. with $c_{e}>0$.

### 9.2.6 Details of the Convergence Process

In the argument to follow, we will encounter the following situation several times. We have a non-negative function $f(t) \geqslant 0$ and we know that $\int_{0}^{\infty} f(t) d t$ is finite. We want to conclude that $f(t)$ converges to zero for $t \rightarrow \infty$. This holds true if $f$ is Lipschitz continuous. Note that the proof of the following lemma is very similar to the proof in [BMV12, Lemma 11]. However, in our case we apply the local Lipschitz condition that we showed in Lemma 9.15.

Lemma 9.27. Let $f(t) \geqslant 0$ for all t. If $\int_{0}^{\infty} f(t) d(t)$ is finite and $f(t)$ is locally Lipschitz continuous, i.e., for every $\varepsilon>0$, there is a $\delta>0$ such that $\left|f\left(t^{\prime}\right)-f(t)\right| \leqslant \varepsilon$ for all $t^{\prime} \in[t, t+\delta]$, then $f(t)$ converges to zero as $t$ goes to infinity. The functions $t \mapsto x^{\mathrm{T}} R|q|-x^{\mathrm{T}} R x=c^{\mathrm{T}}|q|-c^{\mathrm{T}} x$ and $t \mapsto h(t)$ are Lipschitz continuous.

Proof. If $f(t)$ does not converge to zero, there is $\varepsilon>0$ and an infinite unbounded sequence $t_{1}, t_{2}, \ldots$ such that $f\left(t_{i}\right) \geqslant \varepsilon$ for all $i$. Since $f$ is Lipschitz continuous there is $\delta>0$ such that $f\left(t_{i}^{\prime}\right) \geqslant \varepsilon / 2$ for $t_{i}^{\prime} \in\left[t_{i}, t_{i}+\delta\right]$ and all $i$. Hence, the integral $\int_{0}^{\infty} f(t) d t$ is unbounded.

Since $\dot{x}_{e}$ is continuous and bounded (by Lemma 9.16), $x_{e}$ is Lipschitz continuous. Thus, it is enough to show that $q_{e}$ is Lipschitz continuous for all $e$. Since $q_{Z}$ (recall that $Z=\left\{e: c_{e}=0\right\}$ and $P=[m] \backslash Z$ ) is an affine function of $q_{P}$, it suffices to establish the claim for $e \in P$. So let $e \in P$ be such that $c_{e}>0$. First, we claim that $x_{e}(t+\varepsilon) \leqslant(1+2 K \varepsilon) x_{e}$ for all $\varepsilon \leqslant K / 4$, where $K=8 D^{2}\|b\|_{1}\|c\|_{1} / c_{\text {min }}$. Assume that this is not the case. Let

$$
\varepsilon=\inf \left\{\delta \leqslant 1 / 4 K: x_{e}(t+\delta)>(1+2 K \delta) x_{e}(t)\right\}
$$

then $\varepsilon>0$ (since $\dot{x}_{e}(t) \leqslant K x_{e}(t)$ by Lemma 9.19) and, by continuity, $x_{e}(t+\varepsilon) \geqslant(1+2 K \varepsilon) x_{e}(t)$. There must be $t^{\prime} \in[t, t+\varepsilon]$ such that $\dot{x}_{e}\left(t^{\prime}\right)=2 K x_{e}(t)$. On the other hand,

$$
\dot{x}_{e}\left(t^{\prime}\right) \leqslant K x_{e}\left(t^{\prime}\right) \leqslant K(1+2 K \varepsilon) x_{e}(t) \leqslant K(1+2 K / 4 K) x_{e}(t)<2 K x_{e}(t)
$$

which is a contradiction. Thus, $x_{e}(t+\varepsilon) \leqslant(1+2 K \varepsilon) x_{e}$ for all $\varepsilon \leqslant 1 / 4 K$. Similarly, $x_{e}(t+\varepsilon) \geqslant(1-2 K \varepsilon) x_{e}$. Now, let $\alpha=(1-2 K \varepsilon) x_{e}$ and $\beta=(1+2 K \varepsilon) x_{e}$. Then

$$
\left\|q_{e}(t+\delta)|-| q_{e}(t)\right\| \leqslant M\|x(t+\delta)-x(t)\|_{1} \leqslant M m(4 K \varepsilon) x_{e} \leqslant 8 \varepsilon M m K D\left\|b / \gamma_{A}\right\|_{1}
$$

since $x_{e} \leqslant 2 D\left\|b / \gamma_{A}\right\|_{1}$ for sufficiently large $t$ and where $M$ is as in Lemma 9.15. Since $C_{\star}$ is at least $1 / 2$ for all sufficiently large $t$, the division by $C_{\star}$ and $C_{\star}^{2}$ in the definition of $h(t)$ does not affect the claim.

Lemma 9.28. For all $e \in[m]$ of positive cost, it holds that $\left|x_{e}-\left|q_{e}\right|\right| \rightarrow 0$ as $t$ goes to infinity.
Proof. For a trajectory ultimately running in $\mathcal{X}_{1}$, we showed $\frac{d}{d t} \operatorname{cost}(x) \leqslant x^{\mathrm{T}} R|q|-x^{\mathrm{T}} R x \leqslant 0$ with equality if and only if $x=|q|$. Also, $E(q) \leqslant E(x)$, since $x$ dominates a feasible solution. Furthermore, $x^{\mathrm{T}} R|q|-x^{\mathrm{T}} R x$ goes to zero using Lemma 9.27. Thus

$$
\sum_{e} r_{e}\left(x_{e}-\left|q_{e}\right|\right)^{2}=\sum_{e} r_{e} x_{e}^{2}+\sum_{e} r_{e} q_{e}^{2}-2 \sum_{e} r_{e} x_{e}\left|q_{e}\right| \leqslant 2\left(\sum_{e} r_{e} x_{e}^{2}-\sum_{e} r_{e} x_{e}\left|q_{e}\right|\right)
$$

goes to zero. Next observe that there is a constant $C$ such that $x_{e}(t) \leqslant C$ for all $e$ and $t$ as a result of Lemma 9.16. Also $c_{\text {min }}>0$ and hence $r_{e} \geqslant c_{\min } / C$. Thus $\sum_{e} r_{e}\left(x_{e}-\left|q_{e}\right|\right)^{2} \leqslant \frac{C}{c_{\text {min }}} \cdot \sum_{e}\left(x_{e}-\left|q_{e}\right|\right)^{2}$ and hence $\left|x_{e}-\left|q_{e}\right|\right| \rightarrow 0$ for every $e$ with positive cost. For trajectories outside $\mathcal{X}_{1}$, we argue about $\left|\left|q_{e}\right|-\frac{x}{C_{\star}}\right|$ and use $C_{\star} \rightarrow 1$, namely

$$
\sum_{e} r_{e}\left(\frac{x_{e}}{C_{\star}}-\left|q_{e}\right|\right)^{2} \leqslant 2\left(\sum_{e} r_{e}\left(\frac{x_{e}}{C_{\star}}\right)^{2}-\sum_{e} r_{e} \frac{x_{e}}{C_{\star}}\left|q_{e}\right|\right) \rightarrow 0 .
$$

Note that the above does not say anything about the indices $e \in Z$ (with $c_{e}=0$ ). Recall that $A_{P} q_{P}+A_{Z} q_{Z}=b$ and that the columns of $A_{Z}$ are independent. Thus, $q_{Z}$ is uniquely determined by $q_{P}$. For the undirected shortest path problem, the potential difference $p^{\mathrm{T}} b$ between source and sink converges to the length of a shortest source-sink path. If an edge with positive cost is used by some
shortest undirected path, then no shortest undirected path uses it with the opposite direction. We prove the natural generalizations.

Let $\mathcal{S}_{\mathrm{OPT}}$ be the set of optimal solutions to (8.2) and let $E_{\mathrm{OPT}}=\cup_{x \in \mathcal{S}_{\mathrm{OPT}}} \operatorname{supp}(x)$ be the set of columns used in some optimal solution. The columns of positive cost in $E_{\mathrm{OPT}}$ can be consistently oriented as the following Lemma shows.

Lemma 9.29. Let $x_{1}^{*}$ and $x_{2}^{*}$ be optimal solutions to (8.2) and let $f$ and $g$ be feasible solutions with $|f|=x_{1}^{*}$ and $|g|=x_{2}^{*}$. Then there is no $e$ such that $f_{e} g_{e}<0$ and $c_{e}>0$.

Proof. Assume otherwise. Then $\left|g_{e}-f_{e}\right|=\left|g_{e}\right|+\left|f_{e}\right|>0$. Consider $h=\left(g_{e} f-f_{e} g\right) /\left(g_{e}-f_{e}\right)$. Then $A h=\left(g_{e} A f-f_{e} A g\right) /\left(g_{e}-f_{e}\right)=b$ and $h$ is feasible. Also, $h_{e}=\frac{g_{e} f_{e}-f_{e} g_{e}}{g_{e}-f_{e}}=0$ and for every index $e^{\prime}$, it holds that $\left|h_{e^{\prime}}\right|=\frac{\left|g_{e} f_{e^{\prime}}-f_{e} g_{e^{\prime}}\right|}{\left|g_{e}-f_{e}\right| \mid} \leqslant \frac{\left|g_{e}\right|\left|f_{e^{\prime}}\right|+\left|f_{e}\right|\left|g_{e^{\prime}}\right|}{\left|g_{e}\right|+\left|f_{e}\right|}$ and hence

$$
\operatorname{cost}(h)<\operatorname{cost}(f)+\operatorname{cost}(g)=\frac{\left|g_{e}\right|}{\left|g_{e}\right|+\left|f_{e}\right|} \operatorname{cost}\left(x_{1}^{*}\right)+\frac{\left|f_{e}\right|}{\left|g_{e}\right|+\left|f_{e}\right|} \operatorname{cost}\left(x_{2}^{*}\right)=\operatorname{cost}\left(x_{1}^{*}\right),
$$

a contradiction to the optimality of $x_{1}^{*}$ and $x_{2}^{*}$.
By the preceding lemma, we can orient $A$ such that $f_{e} \geqslant 0$ whenever $|f|$ is an optimal solution to (8.2) and $c_{e}>0$. We then call $A$ positively oriented.

Lemma 9.30. It holds that $p^{\mathrm{T}} b$ converges to the cost of an optimum solution of (8.2). If $A$ is positively oriented, then $\liminf _{t \rightarrow \infty} A_{e}^{\mathrm{T}} p \geqslant 0$ for all $e$.

Proof. Let $x^{*}$ be an optimal solution of (8.2). We first show convergence to a point in $L$ and then convergence to $c^{\mathrm{T}} x^{*}$. Let $\varepsilon>0$ be arbitrary. Consider any time $t \geqslant t_{0}$, where $t_{0}$ and $C$ as in Lemma 9.19 and moreover $\left|\left|q_{e}\right|-x_{e}\right| \leqslant \frac{C \varepsilon}{c_{\max }}$ for every $e \in P$. Then $x_{e} \geqslant C$ for all indices $e$ in the support of some basic feasible solution $f$. For every $e \in P$, we have $q_{e}=\frac{x_{e}}{c_{e}} A_{e}^{\mathrm{T}} p$. We also assume $q \geqslant 0$ by possibly reorienting columns of $A$. Hence

$$
\left.\left|c_{e}-A_{e}^{\mathrm{T}} p\right|=\left|1-\frac{q_{e}}{x_{e}}\right| \cdot\left|c_{e}=\left|\frac{x_{e}-q_{e}}{x_{e}}\right| \cdot c_{e} \leqslant \frac{c_{\max }}{C}\right| q_{e}-x_{e} \right\rvert\, \leqslant \varepsilon
$$

For indices $e \in Z$, we have $A_{e}^{\mathrm{T}} p=0=c_{e}$. Since $\|f\|_{\infty} \leqslant D\left\|b / \gamma_{A}\right\|_{1}$ (Lemma 9.10), we conclude

$$
c^{\mathrm{T}} f-p^{\mathrm{T}} b=\sum_{e \in \operatorname{supp}(f)}\left(c_{e}-p^{\mathrm{T}} A_{e}\right) f_{e} \leqslant \varepsilon \sum_{e}\left|f_{e}\right| \leqslant \varepsilon \cdot m D\left\|b / \gamma_{A}\right\|_{1} .
$$

Since the set $L$ is finite, we can let $\varepsilon>0$ be smaller than half the minimal distance between elements in $L$. By the preceding paragraph, there is for all sufficiently large $t$, a basic feasible solution $f$ such that $\left|c^{\mathrm{T}} f-b^{\mathrm{T}} p\right| \leqslant \varepsilon$. Since $b^{\mathrm{T}} p$ is a continuous function of time, $c^{\mathrm{T}} f$ must become constant. We have now shown that $b^{\mathrm{T}} p$ converges to an element in $L$. We will next show that $b^{\mathrm{T}} p$ converges to the optimum cost. Let $x^{*}$ be an optimum solution to (8.2) and let $W=\sum_{e} x_{e}^{*} c_{e} \ln x_{e}$. Since $x(t)$ is bounded, $W$ is bounded. We assume that $A$ is positively oriented, thus there is a feasible $f^{*}$ with $\left|f^{*}\right|=x^{*}$ and $f_{e}^{*} \geqslant 0$ whenever $c_{e}>0$. By reorienting zero cost columns, we may assume $f_{e}^{*} \geqslant 0$ for all $e$. Then $A x^{*}=b$. We have

$$
\begin{aligned}
\dot{W} & =\sum_{e} x_{e}^{*} c_{e} \frac{\left|q_{e}\right|-x_{e}}{x_{e}} \\
& =\sum_{e ; c_{e}>0} x_{e}^{*}\left|A_{e}^{\mathrm{T}} p\right|-\operatorname{cost}\left(x^{*}\right) \\
& =\sum_{e} x_{e}^{*}\left|A_{e}^{\mathrm{T}} p\right|-\operatorname{cost}\left(x^{*}\right) \\
& =\sum_{e} x_{e}^{*}\left(\left|A_{e}^{\mathrm{T}} p\right|-A_{e}^{\mathrm{T}} p\right)+b^{\mathrm{T}} p-\operatorname{cost}\left(x^{*}\right)
\end{aligned}
$$

$$
=\sum_{e ; c_{e}>0} x_{e}^{*}\left|A_{e}^{\mathrm{T}} p\right|-\operatorname{cost}\left(x^{*}\right) \quad \text { since } q_{e}=\frac{x_{e}}{c_{e}} A_{e}^{\mathrm{T}} p \text { whenever } c_{e}>0
$$

$$
\text { since } A_{e}^{\mathrm{T}} p=0 \text { whenever } c_{e}=0
$$

and hence $b^{\mathrm{T}} p-\operatorname{cost}\left(x^{*}\right)$ must converge to zero; note that $b^{\mathrm{T}} p$ is Lipschitz continuous in $t$.
Similarly, $\left|A_{e}^{\mathrm{T}} p\right|-A_{e}^{\mathrm{T}} p$ must converge to zero whenever $x_{e}^{*}>0$. This implies liminf $A_{e}^{\mathrm{T}} p \geqslant 0$. Assume otherwise, i.e., for every $\varepsilon>0$, we have $A_{e}^{\mathrm{T}} p<-\varepsilon$ for arbitrarily large $t$. Since $p$ is Lipschitz continuous in $t$, there is a $\delta>0$ such that $A_{e}^{\mathrm{T}} p<-\varepsilon / 2$ for infinitely many disjoint intervals of length $\delta$. In these intervals, $\left|A_{e}^{\mathrm{T}} p\right|-A_{e}^{\mathrm{T}} p \geqslant \varepsilon$ and hence $W$ must grow beyond any bound, a contradiction.

Corollary 9.31. $E(x)$ and $\operatorname{cost}(x)$ converge to $c^{\mathrm{T}} x^{*}$, whereas $x$ and $|q|$ converge to $\mathcal{S}_{\mathrm{OPT}}$. If the optimum solution is unique, $x$ and $|q|$ converge to it. Moreover, if $e \notin E_{\mathrm{OPT}}, x_{e}$ and $\left|q_{e}\right|$ converge to zero.

Proof. The first part follows from $E(x)=\operatorname{cost}(x)=b^{\mathrm{T}} p$ and the preceding Lemma. Thus $x$ and $q$ converge to the set $F$ of equilibrium points, see (9.12), that are optimum solutions to (8.2). Since every optimum solution is an equilibrium point by Theorem $9.23, x$ and $q$ converge to $\mathcal{S}_{\mathrm{OPT}}$. For $e \notin E_{\mathrm{OPT}}$, $f_{e}=0$ for every $f \in F \cap \mathcal{S}_{\mathrm{OPT}}$. Since $x$ and $|q|$ converge to $F \cap \mathcal{S}_{\mathrm{OPT}}, x_{e}$ and $\left|q_{e}\right|$ converge to zero for every $e \in E_{\text {OPT }}$.

## Chapter 10

## Physarum-Inspired Dynamics

In this chapter, we present in its full generality our main technical result on the Physarum-inspired dynamics (8.6).

### 10.1 Overview

Inspired by the max-flow min-cut theorem, we consider the following primal-dual pair of linear programs: the primal LP is given by $\max \{t: A f=t \cdot b ; 0 \leqslant f \leqslant x\}$ in variables $f \in \mathbb{R}^{m}$ and $t \in \mathbb{R}$, and its dual LP reads $\min \left\{x^{\mathrm{T}} z: z \geqslant 0 ; z \geqslant A^{\mathrm{T}} y ; b^{\mathrm{T}} y=1\right\}$ in variables $z \in \mathbb{R}^{m}$ and $y \in \mathbb{R}^{n}$. Since the dual feasible region does not contain a line and the minimum is bounded, the optimum is attained at a vertex, and in an optimum solution we have $z=\max \left\{0, A^{\mathrm{T}} y\right\}$. Let $V$ be the set of vertices of the dual feasible region, and let $Y \stackrel{\text { def }}{=}\{y:(z, y) \in V\}$ be the set of their projections on $y$-space. Then, the dual optimum is given by $\min \left\{\max \left\{0, y^{\mathrm{T}} A\right\} \cdot x: y \in Y\right\}$. The set of strongly dominating capacity vectors $x$ is defined as

$$
X \stackrel{\text { def }}{=}\left\{x \in \mathbb{R}_{>0}^{m}: y^{\mathrm{T}} A x>0 \text { for all } y \in Y\right\} \cdot .^{1}
$$

Note that $X$ contains the set of all scaled feasible solutions $\{x=t f: A f=b, f \geqslant 0, t>0\}$.
We next discuss the choice of step size. For $y \in Y$ and capacity vector $x$, let $\alpha(y, x) \stackrel{\text { def }}{=} y^{\mathrm{T}} A x$. Further, let $\alpha(x) \stackrel{\text { def }}{=} \min \{\alpha(y, x): y \in Y\}$ and $\alpha_{\ell} \stackrel{\text { def }}{=} \alpha\left(x^{(\ell)}\right)$. Then, for any $x \in X$ there is a feasible $f$ such that $0 \leqslant f \leqslant x / \alpha(x)$, see Lemma 10.8. In particular, if $x$ is feasible then $\alpha(x)=1$, since $\alpha(y, x)=1$ for all $y \in Y$. We partition the Physarum-inspired dynamics (8.6) into the following five regimes and define for each regime a fixed step size, see Section 10.3.
Corollary 10.1. The Physarum-inspired dynamics (8.6) initialized with $x^{(0)} \in X$ and a step size $h$ satisfies:
(i) If $\alpha^{(0)}=1$, we work with $h \leqslant h_{0}$ and have $\alpha_{\ell}=1$ for all $\ell$.
(ii) If $1 / 2 \leqslant \alpha^{(0)}<1$, we work with $h \leqslant h_{0} / 2$ and have $1-\delta \leqslant \alpha_{\ell}<1$ for $\ell \geqslant h^{-1} \log (1 / 2 \delta)$ and $\delta>0$.
(iii) If $1<\alpha^{(0)} \leqslant 1 / h_{0}$, we work with $h \leqslant h_{0}$ and have $1<\alpha_{\ell} \leqslant 1+\delta$ for $\ell \geqslant h^{-1} \cdot \log \left(1 / \delta h_{0}\right)$ and $\delta>0$.
(iv) If $0<\alpha^{(0)}<1 / 2$, we work with $h \leqslant \alpha^{(0)} h_{0}$ and have $1 / 2 \leqslant \alpha_{\ell}<1$ for $\ell \geqslant 1 / h$.
(v) If $1 / h_{0}<\alpha^{(0)}$, we work with $h \leqslant 1 / 4$ and have $1<\alpha_{\ell} \leqslant 1 / h_{0}$ for $\ell=\left\lfloor\log _{1 /(1-h)} h_{0}\left(\alpha^{(0)}-1\right) /(1-\right.$ $\left.\left.h_{0}\right)\right\rfloor$.
In each regime, we have $1-\alpha^{(\ell+1)}=(1-h)\left(1-\alpha_{\ell}\right)$.
We give now the full version of Theorem 8.4 which applies for any strongly dominating starting point.
Theorem 10.2. Suppose $A \in \mathbb{Z}^{n \times m}$ has full row rank $(n \leqslant m), b \in \mathbb{Z}^{n}, c \in \mathbb{Z}_{>0}^{m}$ and $\varepsilon \in(0,1)$. Given $x^{(0)} \in X$ and its corresponding $\alpha_{0}$, the Physarum-inspired dynamics (8.6) initialized with $x^{(0)}$ runs in two regimes:
(i) The first regime is executed when $\alpha^{(0)} \notin\left[1 / 2,1 / h_{0}\right]$ and it computes a point $x^{(t)} \in X$ such that $\alpha_{t} \in\left[1 / 2,1 / h_{0}\right]$. In particular, if $\alpha^{(0)}<1 / 2$ then $h \leqslant(\Phi / \mathrm{opt}) \cdot\left(\alpha_{0} h_{0}\right)^{2}$ and $t=1 / h$. Otherwise, if $\alpha^{(0)}>1 / h_{0}$ then $h \leqslant \Phi /$ opt and $t=\left\lfloor\log _{1 /(1-h)}\left[h_{0}\left(\alpha^{(0)}-1\right) /\left(1-h_{0}\right)\right]\right\rfloor$.
(ii) The second regime starts from a point $x^{(t)} \in X$ with $\alpha_{t} \in\left[1 / 2,1 / h_{0}\right]$, it has a step size $h \leqslant$ ( $\Phi / \mathrm{opt}) \cdot h_{0}^{2} / 2$ and outputs for any $k \geqslant 4 C_{1} /(h \Phi) \cdot \ln \left(C_{2} \Psi^{(0)} /\left(\varepsilon \cdot \min \left\{1, x_{\min }^{(0)}\right\}\right)\right)$ a vector $x^{(t+k)} \in X$ such that $\operatorname{dist}\left(x^{(t+k)}, X_{\star}\right)<\varepsilon /\left(D \gamma_{A}\right)$.
We stated the bounds on $h$ in terms of the unknown quantities $\Phi$ and opt. However, $\Phi /$ opt $\geqslant 1 / C_{3}$ by Lemma 9.10 and hence replacing $\Phi /$ opt by $1 / C_{3}$ yields constructive bounds for $h$.

[^15]Organization: This chapter is devoted to proving Theorem 10.2. It is organized as follows: Section 10.2 establishes core efficiency bounds that extend [SV16c] and yield a scale-invariant determinant dependence of the step size and are applicable to strongly dominating points. Section 10.3 gives the definition of strongly dominating points and shows that the Physarum-inspired dynamics (8.6) initialized with such a point is well defined. Section 10.4 extends the analysis in $\left[\mathrm{BBD}^{+} 13\right.$, SV16b, SV16c $]$ to positive linear programs, by generalizing the concept of non-negative flows to non-negative feasible kernel-free vectors. Section 10.5 shows that $x^{(\ell)}$ converges to $X_{\star}$ for large enough $\ell$. Section 10.6 concludes the proof of Theorem 10.2.

### 10.2 Useful Lemmas

Recall that $R^{(\ell)}=\operatorname{diag}(c) \cdot\left(X^{(\ell)}\right)^{-1}$ is a positive diagonal matrix and $L^{(\ell)} \stackrel{\text { def }}{=} A\left(R^{(\ell)}\right)^{-1} A^{\mathrm{T}}$ is invertible. Let $p^{(\ell)}$ be the unique solution of $L^{(\ell)} p^{(\ell)}=b$. We improve the dependence on $D_{S}$ in [SV16c, Lemma 5.2 ] to $D$.

Lemma 10.3. [SV16c, extension of Lemma 5.2] Suppose $x^{(\ell)}>0, R^{(\ell)}$ is a positive diagonal matrix and $L=A\left(R^{(\ell)}\right)^{-1} A^{\mathrm{T}}$. Then for every $e \in[m]$, it holds that $\left\|A^{\mathrm{T}}\left(L^{(\ell)}\right)^{-1} A_{e}\right\|_{\infty} \leqslant D \cdot c_{e} / x_{e}^{(\ell)}$.

Proof. The statement follows by combining the proof in [SV16c, Lemma 5.2] with Lemma 9.10.
We show next that [SV16b, Corollary 5.3] holds for $x$-capacitated vectors, which extends the class of feasible starting points, and further yields a bound in terms of $D$.

Lemma 10.4. [SV16b, extension of Corollary 5.3] Let $p^{(\ell)}$ be the unique solution of $L^{(\ell)} p^{(\ell)}=b$ and assume $x^{(\ell)}$ is a positive vector with corresponding positive scalar $\alpha_{\ell}$ such that there is a vector $f$ satisfying $A f=\alpha_{\ell} \cdot b$ and $0 \leqslant f \leqslant x^{(\ell)}$. Then $\left\|A^{\mathrm{T}} p^{(\ell)}\right\|_{\infty} \leqslant D\|c\|_{1} / \alpha_{\ell}$.

Proof. By assumption, $f$ satisfies $\alpha_{\ell} b=A f=\sum_{e} f_{e} A_{e}$ and $0 \leqslant f \leqslant x^{(\ell)}$. This yields

$$
\begin{aligned}
\alpha_{\ell}\left\|A^{\mathrm{T}} p^{(\ell)}\right\|_{\infty} & =\left\|A^{\mathrm{T}}\left(L^{(\ell)}\right)^{-1} \cdot \alpha_{\ell} b\right\|_{\infty}=\left\|\sum_{e} f_{e} A^{\mathrm{T}}\left(L^{(\ell)}\right)^{-1} A_{e}\right\|_{\infty} \\
& \leqslant \sum_{e} f_{e}\left\|A^{\mathrm{T}}\left(L^{(\ell)}\right)^{-1} A_{e}\right\|_{\infty} \stackrel{(\text { Lem. 10.3) }}{\leqslant} D \sum_{e} f_{e} \frac{c_{e}}{x_{e}^{(\ell)}} \leqslant D\|c\|_{1} .
\end{aligned}
$$

We note that applying Lemma 10.3 and Lemma 10.4 into the analysis of [SV16c, Theorem 1.3] yields an improved result that depends on the scale-invariant determinant $D$. Moreover, we show in the next Section 10.3 that the Physarum-inspired dynamics (8.6) can be initialized with any strongly dominating point.

We establish now an upper bound on $q$ that does not depend on $x$. We then use this upper bound on $q$ to establish a uniform upper bound on $x$.

Lemma 10.5. For any $x^{(\ell)}>0,\left\|q^{(\ell)}\right\|_{\infty} \leqslant m D^{2}\left\|b / \gamma_{A}\right\|_{1}$.
Proof. Let $f$ be a basic feasible solution of $A f=b$. By definition, $q_{e}^{(\ell)}=\left(x_{e}^{(\ell)} / c_{e}\right) A_{e}^{\mathrm{T}}\left(L^{(\ell)}\right)^{-1} b$ and thus

$$
\left|q_{e}^{(\ell)}\right|=\left|\frac{x_{e}^{(\ell)}}{c_{e}} \sum_{u} A_{e}^{\mathrm{T}}\left(L^{(\ell)}\right)^{-1} A_{u} f_{u}\right| \leqslant \frac{x_{e}^{(\ell)}}{c_{e}} \sum_{u}\left|f_{u}\right| \cdot\left|A_{e}^{\mathrm{T}}\left(L^{(\ell)}\right)^{-1} A_{u}\right| \leqslant D\|f\|_{1}
$$

where the last inequality follows by

$$
\left|A_{e}^{\mathrm{T}}\left(L^{(\ell)}\right)^{-1} A_{u}\right|=\left|A_{u}^{\mathrm{T}}\left(L^{(\ell)}\right)^{-1} A_{e}\right| \leqslant\left\|A^{\mathrm{T}}\left(L^{(\ell)}\right)^{-1} A_{e}\right\|_{\infty} \stackrel{(\mathrm{Lem.1}}{\leqslant}{ }^{10.3)} D \cdot c_{e} / x_{e}^{(\ell)}
$$

By Cramer's rule and Lemma 9.10, we have $\left|q_{e}^{(\ell)}\right| \leqslant D\|f\|_{1} \leqslant m D^{2}\left\|b / \gamma_{A}\right\|_{1}$.
Let $k, t \in \mathbb{N}$. We denote by

$$
\begin{equation*}
\bar{q}^{(t, k)}=\sum_{i=t}^{t+k-1} \frac{h(1-h)^{t+k-1-i}}{1-(1-h)^{k}} q^{(i)} \quad \text { and } \quad \bar{p}^{(t, k)}=\sum_{i=t}^{t+k-1} p^{(i)} . \tag{10.1}
\end{equation*}
$$

Straightforward checking shows that $A \bar{q}^{(t, k)}=b$. Further, for $C \stackrel{\text { def }}{=} \operatorname{diag}(c), t \geqslant 0$ and $k \geqslant 1$, we have

$$
x^{(t)} \prod_{i=t}^{t+k-1}\left[1+h\left(C^{-1} A^{\mathrm{T}} p^{(i)}-1\right)\right]=x^{(t+k)}=(1-h)^{k} x^{(t)}+\left[1-(1-h)^{k}\right] \bar{q}^{(t, k)}
$$

We give next an upper bound on $x^{(k)}$ that is independent of $k$.
Lemma 10.6. Let $\Psi^{(0)}=\max \left\{m D^{2}\left\|b / \gamma_{A}\right\|_{1},\left\|x^{(0)}\right\|_{\infty}\right\}$. Then $\left\|x^{(k)}\right\|_{\infty} \leqslant \Psi^{(0)}, \forall k \in \mathbb{N}$.
Proof. We prove the statement by induction. The base case $\left\|x^{(0)}\right\|_{\infty} \leqslant \Psi^{(0)}$ is clear. Suppose the statement holds for some $k>0$. Then, triangle inequality and Lemma 10.5 yield

$$
\left\|x^{(k+1)}\right\|_{\infty} \leqslant(1-h)\left\|x^{(k)}\right\|_{\infty}+h\left\|q^{(k)}\right\|_{\infty} \leqslant(1-h) \Psi^{(0)}+h \Psi^{(0)} \leqslant \Psi^{(0)}
$$

We show now convergence to feasibility.
Lemma 10.7. Let $r^{(k)}=b-A x^{(k)}$. Then $r^{(k+1)}=(1-h) r^{(k)}$ and hence $r^{(k)}=(1-h)^{k}\left(b-A x^{(0)}\right)$.
Proof. By definition $x^{(k+1)}=(1-h) x^{(k)}+h q^{(k)}$, and thus the statement follows by

$$
r^{(k+1)}=b-A x^{(k+1)}=b-(1-h) A x^{(k)}-h b=(1-h) r^{(k)} .
$$

### 10.3 Strongly Dominating Capacity Vectors

For the shortest path problem, it is known that one can start from any capacity vector $x$ for which the directed capacity of every source-sink cut is positive, where the directed capacity of a cut is the total capacity of the edges crossing the cut in source-sink direction minus the total capacity of the edges crossing the cut in the sink-source direction. We generalize this result. We start with the max-flow like LP

$$
\begin{equation*}
\max \{t: A f=t \cdot b ; 0 \leqslant f \leqslant x\} \tag{10.2}
\end{equation*}
$$

in variables $f \in \mathbb{R}^{m}$ and $t \in \mathbb{R}$ and its dual

$$
\begin{equation*}
\min \left\{x^{\mathrm{T}} z: z \geqslant 0 ; z \geqslant A^{\mathrm{T}} y ; b^{\mathrm{T}} y=1\right\} \tag{10.3}
\end{equation*}
$$

in variables $z \in \mathbb{R}^{m}$ and $y \in \mathbb{R}^{n}$. The feasible region of the dual contains no line. Assume otherwise; say it contains $(z, y)=\left(z^{(0)}, y^{(0)}\right)+\lambda\left(z^{(1)}, y^{(1)}\right)$ for all $\lambda \in \mathbb{R}$. Then, $z \geqslant 0$ implies $z^{(1)}=0$ and further $z \geqslant A^{\mathrm{T}} y$ implies $z^{(0)} \geqslant A^{\mathrm{T}} y^{(0)}+\lambda A^{\mathrm{T}} y^{(1)}$ and hence $A^{\mathrm{T}} y^{(1)}=0$. Since $A$ has full row rank, we have $y^{(1)}=0$. The optimum of the dual is therefore attained at a vertex. In an optimum solution, we have $z=\max \left\{0, A^{\mathrm{T}} y\right\}$. Let $V$ be the set of vertices of the feasible region of the dual (10.3), and let

$$
Y \stackrel{\text { def }}{=}\{y:(z, y) \in V\}
$$

be the set of their projections on $y$-space. Then, the optimum of the dual (10.3) is given by

$$
\begin{equation*}
\min _{y \in Y}\left\{\max \left\{0, y^{\mathrm{T}} A\right\} \cdot x\right\} \tag{10.4}
\end{equation*}
$$

The set of strongly dominating capacity vectors $x$ is defined by

$$
\begin{equation*}
X \stackrel{\text { def }}{=}\left\{x \in \mathbb{R}_{>0}^{m}: y^{\mathrm{T}} A x>0 \text { for all } y \in Y\right\} \tag{10.5}
\end{equation*}
$$

We next show that for all $x^{(0)} \in X$ and sufficiently small step size, the sequence $\left\{x^{(k)}\right\}_{k \in \mathbb{N}}$ stays in $X$. Moreover, $y^{\mathrm{T}} A x^{(k)}$ converges to 1 for every $y \in Y$. We define by

$$
\alpha(y, x) \stackrel{\text { def }}{=} y^{\mathrm{T}} A x \quad \text { and } \quad \alpha(x) \stackrel{\text { def }}{=} \min \{\alpha(y, x): y \in Y\}
$$

Let $\alpha_{\ell} \stackrel{\text { def }}{=} \alpha\left(x^{(\ell)}\right)$. Then, $x^{(\ell)} \in X$ iff $\alpha_{\ell}>0$. We summarize the discussion in the following Lemma.
Lemma 10.8. Suppose $x^{(\ell)} \in X$. Then, there is a vector $f$ such that $A f=\alpha_{\ell} \cdot b$ and $0 \leqslant f \leqslant x^{(\ell)}$.

Proof. By the strong duality theorem applied on (10.2) and (10.3), it holds by (10.4) that

$$
t=\min _{y \in Y}\left\{\max \left\{0, y^{\mathrm{T}} A\right\} \cdot x^{(\ell)}\right\} \geqslant \min _{y \in Y} y^{\mathrm{T}} A x^{(\ell)}=\alpha_{\ell}
$$

The statement follows by the definition of (10.2).
We demonstrate now that $\alpha_{\ell}$ converges to 1 .
Lemma 10.9. Assume $x^{(\ell)} \in X$. Then, for any $h^{(\ell)} \leqslant \min \left\{1 / 4, \alpha_{\ell} h_{0}\right\}$ we have $x^{(\ell+1)} \in X$ and

$$
1-\alpha_{\ell+1}=\left(1-h^{(\ell)}\right) \cdot\left(1-\alpha_{\ell}\right)
$$

Proof. By applying Lemma 10.4 and Lemma 10.8 with $x^{(\ell)} \in X$, we have $\left\|A^{\mathrm{T}} p^{(\ell)}\right\|_{\infty} \leqslant D\|c\|_{1} / \alpha_{\ell}$ and hence for every index $e$ it holds $-h^{(\ell)} \cdot c_{e}^{-1} x_{e}^{(\ell)} A_{e}^{\mathrm{T}} p^{(\ell)} \geqslant-\left(h^{(\ell)} x_{e}^{(\ell)}\right) /\left(2 \alpha_{\ell} h_{0}\right) \geqslant-x_{e}^{(\ell)} / 2$. Thus,

$$
x_{e}^{(\ell+1)}=\left(1-h^{(\ell)}\right) x_{e}^{(\ell)}+h^{(\ell)} \cdot\left[R_{e}^{(\ell)}\right]^{-1} A_{e}^{\mathrm{T}} p^{(\ell)} \geqslant \frac{3}{4} x_{e}^{(\ell)}-\frac{1}{2} x_{e}^{(\ell)}=\frac{1}{4} x_{e}^{(\ell)}>0
$$

Let $y \in Y$ be arbitrary. Then $y^{\mathrm{T}} b=1$ and hence $y^{\mathrm{T}} r^{(\ell)}=y^{\mathrm{T}}\left(b-A x^{(\ell)}\right)=1-y^{\mathrm{T}} A x^{(\ell)}=1-\alpha\left(y, x^{(\ell)}\right)$. The second claim now follows from Lemma 10.7.

We note that the convergence speed crucially depends on the initial point $x^{(0)} \in X$, and in particular to its corresponding value $\alpha_{0}$. Further, this dependence naturally partitions the Physarum-inspired dynamics (8.6) into the five regimes given in Corollary 10.1.

## $10.4 x^{(k)}$ is Close to a Non-Negative Kernel-Free Vector

In this section, we generalize [SV16b, Lemma 5.4] to positive linear programs. We achieve this in two steps. First, we generalize a result by Ito et al. [IJNT11, Lemma 2] to positive linear programs and then we substitute the notion of a non-negative cycle-free flow with a non-negative feasible kernel-free vector.

Throughout this and the consecutive section, we denote by $\rho_{A} \stackrel{\text { def }}{=} \max \left\{D \gamma_{A}, n D^{2}\|A\|_{\infty}\right\}$.
Lemma 10.10. Suppose a matrix $A \in \mathbb{Z}^{n \times m}$ has full row rank and vector $b \in \mathbb{Z}^{n}$. Let $g$ be a feasible solution to $A g=b$ and $S \subseteq[n]$ be a subset of row indices of $A$ such that $\sum_{i \in S}\left|g_{i}\right|<1 / \rho_{A}$. Then, there is a feasible solution $f$ such that $g_{i} \cdot f_{i} \geqslant 0$ for all $i \in[n], f_{i}=0$ for all $i \in S$ and $\|f-g\|_{\infty}<1 /\left(D \gamma_{A}\right)$.
Proof. W.l.o.g. we can assume that $g \geqslant 0$ as we could change the signs of the columns of $A$ accordingly. Let $\mathbf{1}_{S}$ be the indicator vector of $S$. We consider the linear program

$$
\min \left\{\mathbf{1}_{S}^{\mathrm{T}} x: A x=b, x \geqslant 0\right\}
$$

and let opt be its optimum value. Notice that $0 \leqslant o p t \leqslant \mathbf{1}_{S}^{\mathrm{T}} g<1 / \rho_{A}$. Since the feasible region does not contain a line and the minimum is bounded, the optimum is attained at a basic feasible solution, say $f$. Suppose that there is an index $i \in S$ with $f_{i}>0$. By Lemma 9.10, we have $f_{i} \geqslant 1 /\left(D \gamma_{A}\right)$. This is a contradiction to the optimality of $f$ and hence $f_{i}=0$ for all $i \in S$.

Among the feasible solutions $f$ such that $f_{i} g_{i} \geqslant 0$ for all $i$ and $f_{i}=0$ for all $i \in S$, we choose the one that minimizes $\|f-g\|_{\infty}$. For simplicity, we also denote it by $f$. Note that $f$ satisfies $\operatorname{supp}(f) \subseteq \bar{S}$, where $\bar{S}=[m] \backslash S$. Further, since $f_{S}=0$ and

$$
A_{S} g_{S}+A_{\bar{S}} g_{\bar{S}}=A g=b=A f=A_{S} f_{S}+A_{\bar{S}} f_{\bar{S}}=A_{\bar{S}} f_{\bar{S}}
$$

we have $A_{\bar{S}}\left(f_{\bar{S}}-g_{\bar{S}}\right)=A_{S} g_{S}$. Let $A_{B}$ be a linearly independent column subset of $A_{\bar{S}}$ with maximal cardinality, i.e. the column subset $A_{N}$, where $N=\bar{S} \backslash B$, is linearly dependent on $A_{B}$. Hence, there is an invertible square submatrix $A_{B}^{\prime} \in \mathbb{Z}^{|B| \times|B|}$ of $A_{B}$ and a vector $v=\left(v_{B}, 0_{N}\right)$ such that

$$
\binom{A_{B}^{\prime}}{A_{B}^{\prime \prime}} v_{B}=A_{B} v_{B}=A_{S} g_{S} .
$$

Let $r=\left(A_{S} g_{S}\right)_{B}$. Since $A_{B}^{\prime}$ is invertible, there is a unique vector $v_{B}$ such that $A_{B}^{\prime} v_{B}=r$. Observe that

$$
\left|r_{i}\right|=\left|\sum_{j \in S} A_{i, j} g_{j}\right| \leqslant\|A\|_{\infty} \sum_{j \in S}\left|g_{j}\right|<\frac{\|A\|_{\infty}}{n D^{2}\|A\|_{\infty}}=\frac{1}{n D^{2}}
$$

By Cramer's rule $v_{B}(e)$ is quotient of two determinants. The denominator is $\operatorname{det}\left(A_{B}^{\prime}\right)$ and hence at least one in absolute value. For the numerator, the $e$-th column is replaced by $r$. Expansion according to this column shows that the absolute value of the numerator is bounded by

$$
\frac{D}{\gamma_{A}} \sum_{i \in B}\left|r_{i}\right|<\frac{D}{\gamma_{A}} \cdot \frac{|B|}{n D^{2}} \leqslant \frac{1}{D \gamma_{A}}
$$

Therefore, $\|f-g\|_{\infty} \leqslant 1 /\left(D \gamma_{A}\right)$ and the statement follows.
Lemma 10.11. Let $q \in \mathbb{R}^{m}, p \in \mathbb{R}^{n}$ and $N=\left\{e \in[m]: q_{e} \leqslant 0\right.$ or $\left.p^{T} A_{e} \leqslant 0\right\}$, where $A q=b$ and $p=L^{-1} b$. Suppose $\sum_{e \in N}\left|q_{e}\right|<1 / \rho_{A}$. Then there is a non-negative feasible kernel-free vector $f$ such that $\operatorname{supp}(f) \subseteq E \backslash N$ and $\|f-q\|_{\infty}<1 /\left(D \gamma_{A}\right)$.

Proof. We apply Lemma 10.10 to $q$ with $S=N$. Then, there is a non-negative feasible vector $f$ such that $\operatorname{supp}(f) \subseteq E \backslash N$ and $\|f-q\|_{\infty}<1 /\left(D \gamma_{A}\right)$. By Lemma $9.1, f$ can be expressed as a sum of a convex combination of basic feasible solutions plus a vector $w$ in the kernel of $A$. Moreover, all vectors in this representation are sign compatible with $f$, and in particular $w$ is non-negative too.

Suppose for contradiction that $w \neq 0$. By definition, $0=p^{\mathrm{T}} A w=\sum_{e \in[m]} p^{\mathrm{T}} A_{e} w_{e}$ and since $w \geqslant 0$ and $w \neq 0$, it follows that there is an index $e \in[m]$ satisfying $w_{e}>0$ and $p^{\mathrm{T}} A_{e} \leqslant 0$. Since $f$ and $w$ are sign compatible, $w_{e}>0$ implies $f_{e}>0$. On the other hand, as $p^{\mathrm{T}} A_{e} \leqslant 0$ we have $e \in N$ and thus $f_{e}=0$. This is a contradiction, hence $w=0$.

Using Corollary 10.1, for any point $x^{(0)} \in X$ there is a point $x^{(t)} \in X$ such that $\alpha_{t} \in\left[1 / 2,1 / h_{0}\right]$. Thus, we can assume that $\alpha_{0} \in\left[1 / 2,1 / h_{0}\right]$ and work with $h \leqslant h_{0} / 2$, where $h_{0}=c_{\min } /\left(2 D\|c\|_{1}\right)$. We generalize next [SV16b, Lemma 5.4].

Lemma 10.12. Suppose $x^{(t)} \in X$ such that $\alpha_{t} \in\left[1 / 2,1 / h_{0}\right], h \leqslant h_{0} / 2$ and $\varepsilon \in(0,1)$. Then, for any $k \geqslant h^{-1} \ln \left(8 m \rho_{A} \Psi^{(0)} / \varepsilon\right)$ there is a non-negative feasible kernel-free vector $f$ such that $\left\|x^{(t+k)}-f\right\|_{\infty}<$ $\varepsilon /\left(D \gamma_{A}\right)$.

Proof. Let $\beta^{(k)} \stackrel{\text { def }}{=} 1-(1-h)^{k}$. By (10.1), vector $\bar{q}^{(t, k)}$ satisfies $A \bar{q}^{(t, k)}=b$ and thus Lemma 10.6 yields

$$
\begin{equation*}
\left\|x^{(t+k)}-\beta^{(k)} \bar{q}^{(t, k)}\right\|_{\infty}=(1-h)^{k} \cdot\left\|x^{(t)}\right\|_{\infty} \leqslant \exp \{-h k\} \cdot \Psi^{(0)} \leqslant \varepsilon /\left(8 m \rho_{A}\right) \tag{10.6}
\end{equation*}
$$

Using Corollary 10.1, we have $x^{(t+k)} \in X$ such that $\alpha^{(t+k)} \in\left(1 / 2,1 / h_{0}\right)$ for every $k \in \mathbb{N}_{+}$. Let $F_{k}=$ $Q_{k} \cup P_{k}$, where $Q_{k}=\left\{e \in[m]: \bar{q}_{e}^{(t, k)} \leqslant 0\right\}$ and $P_{k}=\left\{e \in[m]: A_{e}^{\mathrm{T}} \bar{p}^{(t, k)} \leqslant 0\right\}$. Then, for every $e \in Q_{k}$ it holds

$$
\begin{equation*}
\left|\bar{q}_{e}^{(t, k)}\right| \leqslant\left[\beta^{(k)}\right]^{-1} \cdot\left|x_{e}^{(t+k)}-\beta^{(k)} \bar{q}_{e}^{(t, k)}\right| \leqslant \varepsilon /\left(7 m \rho_{A}\right) \tag{10.7}
\end{equation*}
$$

By Lemma $10.6,\left\|x^{(\cdot)}\right\| \leqslant \Psi^{(0)}$. Moreover, by (10.1) for every $e \in P_{k}$ we have

$$
\begin{aligned}
x_{e}^{(t+k)} & =x_{e}^{(t)} \prod_{i=t}^{k+t-1}\left[1+h\left(c_{e}^{-1} A_{e}^{\mathrm{T}} p^{(i)}-1\right)\right] \\
& \leqslant x_{e}^{(t)} \cdot \exp \left\{-h k+\left(h / c_{e}\right) \cdot A_{e}^{\mathrm{T}} \bar{p}^{(t, k)}\right\} \\
& \leqslant \exp \{-h k\} \cdot \Psi^{(0)} \\
& \leqslant \varepsilon /\left(8 m \rho_{A}\right)
\end{aligned}
$$

and by combining the triangle inequality with (10.6), it follows for every $e \in P_{k}$ that

$$
\begin{align*}
\left|\bar{q}_{e}^{(t, k)}\right| & \leqslant\left[\beta^{(k)}\right]^{-1} \cdot\left[\left|x_{e}^{(t+k)}-\beta^{(k)} \bar{q}_{e}^{(t, k)}\right|+\left|x_{e}^{(t+k)}\right|\right] \\
& \leqslant\left[\beta^{(k)}\right]^{-1} \cdot \varepsilon /\left(4 m \rho_{A}\right) \\
& \leqslant \varepsilon /\left(3 m \rho_{A}\right) \tag{10.8}
\end{align*}
$$

Therefore, (10.7) and (10.8) yields that

$$
\begin{equation*}
\sum_{e \in F_{k}}\left|\bar{q}_{e}^{(t, k)}\right| \leqslant m \cdot \varepsilon /\left(3 m \rho_{A}\right) \leqslant \varepsilon /\left(3 \rho_{A}\right) \tag{10.9}
\end{equation*}
$$

By Lemma 10.11 applied with $\bar{q}_{e}^{(t, k)}$ and $N=F_{k}$, it follows by (10.9) that there is a non-negative feasible kernel-free vector $f$ such that $\operatorname{supp}(f) \subseteq E \backslash N$ and

$$
\left\|f-\bar{q}^{(t, k)}\right\|_{\infty}<\varepsilon /\left(3 D \gamma_{A}\right)
$$

By Lemma 10.5 , we have $\left\|\bar{q}^{(t, k)}\right\|_{\infty} \leqslant m D^{2}\left\|b / \gamma_{A}\right\|_{1}$ and since $\Psi^{(0)} \geqslant m D^{2}\left\|b / \gamma_{A}\right\|_{1}$, it follows that

$$
\begin{aligned}
\left\|x^{(t+k)}-f\right\|_{\infty} & =\left\|x^{(t+k)}-\beta^{(k)} \bar{q}^{(t+k)}+\beta^{(k)} \bar{q}^{(t+k)}-f\right\|_{\infty} \\
& \leqslant\left\|x^{(t+k)}-\beta^{(k)} \bar{q}^{(t+k)}\right\|_{\infty}+\left\|f-\bar{q}^{(t+k)}\right\|_{\infty}+(1-h)^{k}\left\|\bar{q}^{(t+k)}\right\|_{\infty} \\
& \leqslant \frac{\varepsilon}{8 m \rho_{A}}+\frac{\varepsilon}{3 D \gamma_{A}}+\frac{\varepsilon \cdot m D^{2}\left\|b / \gamma_{A}\right\|_{1}}{8 m \rho_{A} \cdot \Psi^{(0)}} \leqslant \frac{\varepsilon}{D \gamma_{A}} .
\end{aligned}
$$

## $10.5 x^{(k)}$ is $\varepsilon$-Close to an Optimal Solution

Recall that $\mathcal{N}$ denotes the set of non-optimal basic feasible solutions of (8.5) and $\Phi=\min _{g \in \mathcal{N}} c^{\mathrm{T}} g-\mathrm{opt}$. For completeness, we prove next a well known inequality [PS82, Lemma 8.6] that lower bounds the value of $\Phi$.

Lemma 10.13. Suppose $A \in \mathbb{R}^{n \times m}$ has full row rank, $b \in \mathbb{R}^{n}$ and $c \in \mathbb{R}^{m}$ are integral. Then, $\Phi \geqslant$ $1 /\left(D \gamma_{A}\right)^{2}$.

Proof. Let $g=\left(g_{B}, 0\right)$ be an arbitrary basic feasible solution with basis matrix $A_{B}$, where $g_{B}(e) \neq 0$ and $\left|\operatorname{supp}\left(g_{B}\right)\right|=n$. We write $M_{-i,-j}$ to denote the matrix $M$ with deleted $i$-th row and $j$-th column. Let $Q_{e}$ be the matrix formed by replacing the $e$-th column of $A_{B}$ by the column vector $b$. Then, by Cramer's rule, we have

$$
\left|g_{B}(e)\right|=\left|\frac{\operatorname{det}\left(Q_{e}\right)}{\operatorname{det}\left(A_{B}\right)}\right|=\frac{1}{\gamma_{A}}\left|\sum_{k=1}^{n} \frac{(-1)^{j+k} \cdot b_{k} \cdot \operatorname{det}\left(\gamma_{A}^{-1}\left[A_{B}\right]_{-k,-j}\right)}{\operatorname{det}\left(\gamma_{A}^{-1} A_{B}\right)}\right| \geqslant \frac{1}{D \gamma_{A}} .
$$

Note that all components of vector $g_{B}$ have denominator with equal value, i.e. $\operatorname{det}\left(A_{B}\right)$. Consider an arbitrary non-optimal basic feasible solution $g$ and an optimal basic feasible solution $f^{\star}$. Then, $g_{e}=G_{e} / G$ and $f_{e}^{\star}=F_{e} / F$ are rationals such that $G_{e}, G, F_{e}, F \leqslant D \gamma_{A}$ for every $e$. Further, let $r_{e}=c_{e}\left(G_{e} F-F_{e} F\right) \in$ $\mathbb{Z}$ for every $e \in[m]$, and observe that

$$
c^{\mathrm{T}}\left(g-f^{\star}\right)=\sum_{e} c_{e}\left(g_{e}-f_{e}^{\star}\right)=\frac{1}{G F} \sum_{e} r_{e} \geqslant 1 /\left(D \gamma_{A}\right)^{2},
$$

where the last inequality follows by $c^{\mathrm{T}}\left(g-f^{\star}\right)>0$ implies $\sum_{e} r_{e} \geqslant 1$.
Lemma 10.14. Let $f$ be a non-negative feasible kernel-free vector and $\varepsilon \in(0,1)$ a parameter. Suppose for every non-optimal basic feasible solution $g$, there exists an index $e \in[m]$ such that $g_{e}>0$ and $f_{e}<\varepsilon /\left(2 m D^{3} \gamma_{A}\|b\|_{1}\right)$. Then, $\left\|f-f^{\star}\right\|_{\infty}<\varepsilon /\left(D \gamma_{A}\right)$ for some optimal $f^{\star}$.

Proof. Let $C=2 D^{2}\|b\|_{1}$. Since $f$ is kernel-free, by Lemma 9.1 it can be expressed as a convex combination of sign-compatible basic feasible solutions $f=\sum_{i=1}^{\ell} \alpha_{i} f^{(i)}+\sum_{i=\ell+1}^{m} \alpha_{i} f^{(i)}$, where $f^{(1)}, \ldots, f^{(\ell)}$ denote the optimal solutions. By Lemma 9.10, $f_{e}^{(i)}>0$ implies $f_{e}^{(i)} \geqslant 1 /\left(D \gamma_{A}\right)$. By the hypothesis, for every non-optimal $f^{(i)}$, i.e. $i \geqslant \ell+1$, there exists an index $e(i) \in[m]$ such that

$$
1 /\left(D \gamma_{A}\right) \leqslant f_{e(i)}^{(i)} \quad \text { and } \quad f_{e(i)}<\varepsilon /\left(m D \gamma_{A} \cdot C\right)
$$

Therefore, we have

$$
\alpha_{i} /\left(D \gamma_{A}\right) \leqslant \alpha_{i} f_{e(i)}^{(i)} \leqslant \sum_{j=1}^{m} \alpha_{j} f_{e(i)}^{(j)}=f_{e(i)}<\varepsilon /\left(m D \gamma_{A} \cdot C\right),
$$

and hence $\sum_{i=\ell+1}^{m} \alpha_{i} \leqslant \varepsilon / C$. Further, by Lemma 9.10 , for every $j$ we have

$$
\left\|f^{(j)}\right\|_{\infty} \leqslant D\left\|b / \gamma_{A}\right\|_{1}=C /\left(2 D \gamma_{A}\right)
$$

Let $\beta \geqslant 0$ be an arbitrary vector satisfying $\sum_{i=1}^{\ell} \beta_{i}=\sum_{i=\ell+1}^{m} \alpha_{i}$. Let $\nu_{i}=\alpha_{i}+\beta_{i}$ for every $i \in[\ell]$ and let $f^{\star}=\sum_{i=1}^{\ell} \nu_{i} f^{(i)}$. Then, $f^{\star}$ is an optimal solution and we have

$$
\begin{aligned}
\left\|f^{\star}-f\right\|_{\infty} & =\left\|\sum_{i=1}^{\ell} \beta_{i} f^{(i)}-\sum_{i=\ell+1}^{m} \alpha_{i} \cdot f^{(i)}\right\|_{\infty} \\
& \leqslant \max _{i \in[1: m]}\left\|f^{(i)}\right\|_{\infty} \cdot\left(\sum_{i=1}^{\ell} \beta_{i}+\sum_{i=\ell+1}^{m} \alpha_{i}\right) \\
& \leqslant \frac{2 \varepsilon}{C} \cdot \frac{C}{2 D \gamma_{A}}=\frac{\varepsilon}{D \gamma_{A}}
\end{aligned}
$$

In the following lemma, we extend the analysis in [SV16b, Lemma 5.6] from the transshipment problem to positive linear programs. Our result crucially relies on an argument that uses the parameter $\Phi=\min _{g \in \mathcal{N}} c^{\mathrm{T}} g$ - opt. It is here, where our analysis incurs the linear step size dependence on $\Phi /$ opt and the quadratic dependence on opt/ $\Phi$ for the number of steps.

An important technical detail is that the first regime incurs an extra ( $\Phi /$ opt)-factor dependence. At first glance, this might seem unnecessary due to Corollary 10.1, however it is crucial for an inductive argument in our analysis to hold, see (10.11) and (10.12). Further, we note that the undirected Physarum dynamics (8.7) satisfies $x_{\min }^{(t)} \geqslant(1-h)^{t} \cdot x_{\min }^{(0)}$, whereas the directed Physarum-inspired dynamics (8.6) might yield a value $x_{\text {min }}^{(t)}$ which decreases with faster than exponential rate. As our analysis incurs a logarithmic dependence on $1 / x_{\min }^{(0)}$, it is prohibitive to decouple the two regimes and give bounds in terms of $\log \left(1 / x_{\text {min }}^{(t)}\right)$, which would be necessary as $x^{(t)}$ is the initial point of the second regime.
Lemma 10.15. Let $g$ be an arbitrary non-optimal basic feasible solution. Given $x^{(0)} \in X$ and its corresponding $\alpha_{0}$, the Physarum-inspired dynamics (8.6) initialized with $x^{(0)}$ runs in two regimes:

1. The first regime is executed when $\alpha^{(0)} \notin\left[1 / 2,1 / h_{0}\right]$ and computes a point $x^{(t)} \in X$ such that $\alpha_{t} \in\left[1 / 2,1 / h_{0}\right]$. In particular, if $\alpha^{(0)}<1 / 2$ then $h \leqslant(\Phi / \mathrm{opt}) \cdot\left(\alpha_{0} h_{0}\right)^{2}$ and $t=1 / h$. Otherwise, if $\alpha^{(0)}>1 / h_{0}$ then $h \leqslant \Phi /$ opt and $t=\left\lfloor\log _{1 /(1-h)}\left[h_{0}\left(\alpha^{(0)}-1\right) /\left(1-h_{0}\right)\right]\right\rfloor$.
2. The second regime starts from a point $x^{(t)} \in X$ such that $\alpha_{t} \in\left[1 / 2,1 / h_{0}\right]$, it has step size $h \leqslant$ (opt $/ \Phi$ ) $\cdot h_{0}^{2} / 2$ and for any $k \geqslant 4 \cdot c^{\mathrm{T}} g /(h \Phi) \cdot \ln \left(\Psi^{(0)} / \varepsilon x_{\min }^{(0)}\right.$ ), guarantees the existence of an index $e \in[m]$ such that $g_{e}>0$ and $x_{e}^{(t+k)}<\varepsilon$.

Proof. Similar to the work of $\left[\mathrm{BBD}^{+} 13\right.$, SV16b], we use a potential function that takes as input a basic feasible solution $g$ and a step number $\ell$, and is defined by

$$
\mathcal{B}_{g}^{(\ell)} \stackrel{\text { def }}{=} \sum_{e \in[m]} g_{e} c_{e} \ln x_{e}^{(\ell)} .
$$

Since $x_{e}^{(\ell+1)}=x_{e}^{(\ell)}\left(1+h^{(\ell)}\left[c_{e}^{-1} \cdot A_{e}^{\mathrm{T}} p^{(\ell)}-1\right]\right)$, we have

$$
\begin{align*}
\mathcal{B}_{g}^{(\ell+1)}-\mathcal{B}_{g}^{(\ell)} & =\sum_{e} g_{e} c_{e} \ln \frac{x_{e}^{(\ell+1)}}{x_{e}^{(\ell)}}=\sum_{e} g_{e} c_{e} \ln \left(1+h^{(\ell)}\left[\frac{A_{e}^{\mathrm{T}} p^{(\ell)}}{c_{e}}-1\right]\right) \\
& \leqslant h^{(\ell)} \sum_{e} g_{e} c_{e}\left[\frac{A_{e}^{\mathrm{T}} p^{(\ell)}}{c_{e}}-1\right]=h^{(\ell)}\left[-c^{\mathrm{T}} g+\left[p^{(\ell)}\right]^{\mathrm{T}} A g\right]=h^{(\ell)}\left[-c^{\mathrm{T}} g+b^{\mathrm{T}} p^{(\ell)}\right] . \tag{10.10}
\end{align*}
$$

Let $f^{\star}$ be an optimal solution to (8.5). In order to lower bound $\mathcal{B}_{f^{\star}}^{(\ell+1)}-\mathcal{B}_{f^{\star}}^{(\ell)}$, we use the inequality $\ln (1+x) \geqslant x-x^{2}$, for all $x \in\left[-\frac{1}{2}, \frac{1}{2}\right]$. Then, we have

$$
\begin{align*}
\mathcal{B}_{f^{\star}}^{(\ell+1)}-\mathcal{B}_{f^{\star}}^{(\ell)} & =\sum_{e} f_{e}^{\star} c_{e} \ln \frac{x_{e}^{(\ell+1)}}{x_{e}^{(\ell)}}=\sum_{e} f_{e}^{\star} c_{e} \ln \left(1+h^{(\ell)}\left[\frac{A_{e}^{\mathrm{T}} p^{(\ell)}}{c_{e}}-1\right]\right) \\
& \geqslant \sum_{e} f_{e}^{\star} c_{e}\left(h^{(\ell)}\left[\frac{A_{e}^{\mathrm{T}} p^{(\ell)}}{c_{e}}-1\right]-\left[h^{(\ell)}\right]^{2}\left[\frac{A_{e}^{\mathrm{T}} p^{(\ell)}}{c_{e}}-1\right]^{2}\right) \\
& \geqslant h^{(\ell)}\left(b^{\mathrm{T}} p^{(\ell)}-\mathrm{opt}-h^{(\ell)} \cdot\left(1 / 2 \alpha_{\ell} h_{0}\right)^{2} \cdot \mathrm{opt}\right), \tag{10.11}
\end{align*}
$$

where the last inequality follows by combining

$$
\sum_{e} f_{e}^{\star} c_{e}\left[\left(c_{e}^{-1} A_{e}^{\mathrm{T}} p^{(\ell)}\right)-1\right]=\left[p^{(\ell)}\right]^{\mathrm{T}} A f^{\star}-\mathrm{opt}=b^{\mathrm{T}} p^{(\ell)}-\mathrm{opt},
$$

$\left\|A^{\mathrm{T}} p^{(\ell)}\right\|_{\infty} \leqslant D\|c\|_{1} / \alpha_{\ell}$ (by Lemma 10.4 and Lemma 10.8 applied with $\left.x^{(\ell)} \in X\right), h_{0}=c_{\text {min }} /\left(4 D\|c\|_{1}\right)$ and

$$
h^{(\ell)} \sum_{e} f_{e}^{\star} c_{e} \cdot\left(c_{e}^{-1} A_{e}^{\mathrm{T}} p^{(\ell)}-1\right)^{2} \leqslant h^{(\ell)}\left(2 D\|c\|_{1} / \alpha_{\ell} c_{\min }\right)^{2} \mathrm{opt}=h^{(\ell)}\left(1 / 2 \alpha_{\ell} h_{0}\right)^{2} \mathrm{opt}
$$

Further, by combining (10.10), (10.11), $c^{\mathrm{T}} g-$ opt $\geqslant \Phi$ for every non-optimal basic feasible solution $g$ and provided that the inequality $h^{(\ell)}\left(1 / 2 \alpha_{\ell} h_{0}\right)^{2}$ opt $\leqslant \Phi / 2$ holds, we obtain

$$
\begin{equation*}
\mathcal{B}_{f^{\star}}^{(\ell+1)}-\mathcal{B}_{f^{\star}}^{(\ell)} \geqslant h^{(\ell)}\left(b^{\mathrm{T}} p^{(\ell)}-c^{\mathrm{T}} g\right)+h^{(\ell)}\left(c^{\mathrm{T}} g-\mathrm{opt}-\frac{\Phi}{2}\right) \geqslant \mathcal{B}_{g}^{(\ell+1)}-\mathcal{B}_{g}^{(\ell)}+\frac{h^{(\ell)} \Phi}{2} \tag{10.12}
\end{equation*}
$$

Using Corollary 10.1, we partition the Physarum-inspired dynamics (8.6) execution into three regimes, based on $\alpha_{0}$. For every $i \in\{1,2,3\}$, we show next that the $i$-th regime has a fixed step size $h^{(\ell)}=h_{i}$ such that $h^{(\ell)}\left(1 / 2 \alpha_{\ell} h_{0}\right)^{2}$ opt $\leqslant \Phi / 2$, for every step $\ell$ in this regime.

By Lemma 10.9, for every $i \in\{1,2,3\}$ it holds for every step $\ell$ in the $i$-th regime that

$$
\begin{equation*}
\alpha_{\ell}=1-\left(1-h_{i}\right)^{\ell} \cdot\left(1-\alpha_{0}\right) . \tag{10.13}
\end{equation*}
$$

Case 1: Suppose $\alpha_{0}>1 / h_{0}$. Notice that $h^{(\ell)}=\Phi /$ opt suffices, since $1 /\left(2 \alpha_{\ell} h_{0}\right)<1 / 2$ for every $\alpha_{\ell}>1 / h_{0}$. Further, by applying (10.13) with $\alpha_{t} \stackrel{\text { def }}{=} 1 / h_{0}$, we have $t=\left\lfloor\log _{1 /\left(1-h^{(\ell)}\right)}\left[h_{0}\left(\alpha^{(0)}-1\right) /(1-\right.\right.$ $\left.\left.\left.h_{0}\right)\right]\right\rfloor \leqslant(\mathrm{opt} / \Phi) \cdot \log \left(\alpha_{0} h_{0}\right)$. Note that by (10.13) the sequence $\left\{\alpha_{\ell}\right\}_{\ell \leqslant t}$ is decreasing, and by Corollary 10.1 we have $1<\alpha_{t} \leqslant 1 / h_{0}$.

Case 2: Suppose $\alpha_{0} \in(0,1 / 2)$. By (10.13) the sequence $\left\{\alpha_{\ell}\right\}_{\ell \in \mathbb{N}}$ is increasing and by Corollary 10.1 the regime is terminated once $\alpha_{\ell} \in[1 / 2,1)$. Observe that $h^{(\ell)}=(\Phi /$ opt $) \cdot\left(\alpha_{0} h_{0}\right)^{2}$ suffices, since $\alpha_{0} \leqslant \alpha_{\ell}$. Then, by (10.13) applied with $\alpha_{t} \stackrel{\text { def }}{=} 1 / 2$, this regime has at most $t=($ opt $/ \Phi) \cdot\left(1 / \alpha_{0} h_{0}\right)^{2}$ steps.

Case 3: Suppose $\alpha_{0} \in\left[1 / 2,1 / h_{0}\right]$. By (10.13) the sequence $\left\{\alpha_{\ell}\right\}_{\ell \in \mathbb{N}}$ converges to 1 (decreases if $\alpha_{0} \in\left(1,1 / h_{0}\right]$ and increases when $\alpha_{0} \in[1 / 2,1)$. Notice that $h^{(\ell)}=(\Phi /$ opt $) \cdot h_{0}^{2} / 2$ suffices, since $1 / 2 \leqslant$ $\alpha_{\ell} \leqslant 1 / h_{0}$ for every $\ell \in \mathbb{N}$. We note that the number of steps in this regime is to be determined soon.

Hence, we conclude that inequality (10.12) holds. Further, using Case 1 and Case 2 there is an integer $t \in \mathbb{N}$ such that $\alpha_{t} \in\left[1 / 2,1 / h_{0}\right]$. Let $k \in \mathbb{N}$ be the number of steps in Case 3 , and let $h \stackrel{\text { def }}{=}(\Phi / \mathrm{opt}) \cdot h_{0}^{2} / 2$. Then, for every $\ell \in\{t, \ldots, t+k-1\}$ it holds that $h^{(\ell)}=h$ and thus

$$
\begin{equation*}
\mathcal{B}_{f^{\star}}^{(t+k)}-\mathcal{B}_{f^{\star}}^{(0)} \geqslant \mathcal{B}_{g}^{(t+k)}-\mathcal{B}_{g}^{(0)}+\sum_{\ell=0}^{t+k-1} \frac{h^{(\ell)} \Phi}{2} \geqslant \mathcal{B}_{g}^{(t+k)}-\mathcal{B}_{g}^{(0)}+k \cdot \frac{h \Phi}{2} . \tag{10.14}
\end{equation*}
$$

By Lemma $10.6, \mathcal{B}_{g}^{(\ell)} \leqslant c^{\mathrm{T}} g \cdot \ln \Psi^{(0)}$ for every basic feasible solution $g$ and every $\ell \in \mathbb{N}$, and thus

$$
\begin{aligned}
\mathcal{B}_{g}^{(t+k)} & \leqslant-k \cdot \frac{h \Phi}{2}+\mathcal{B}_{g}^{(0)}+\mathcal{B}_{f^{\star}}^{(t+k)}-\mathcal{B}_{f^{\star}}^{(0)} \\
& \leqslant-k \cdot \frac{h \Phi}{2}+c^{\mathrm{T}} g \cdot \ln \Psi^{(0)}+\mathrm{opt} \cdot \ln \Psi^{(0)}-\mathrm{opt} \cdot \ln x_{\min }^{(0)} \\
& \leqslant-k \cdot \frac{h \Phi}{2}+2 c^{\mathrm{T}} g \cdot \ln \frac{\Psi^{(0)}}{x_{\min }^{(0)}}
\end{aligned}
$$

Suppose for the sake of a contradiction that for every $e \in[m]$ with $g_{e}>0$ it holds $x_{e}^{(t+k)}>\varepsilon$. Then, $\mathcal{B}_{g}^{(t+k)}>c^{\mathrm{T}} g \cdot \ln \varepsilon$ yields $k<4 \cdot c^{\mathrm{T}} g /(h \Phi) \cdot \ln \left(\Psi^{(0)} /\left(\varepsilon x_{\text {min }}^{(0)}\right)\right)$, a contradiction to the choice of $k$.

### 10.6 Proof of Theorem 10.2

By Corollary 10.1 and Lemma 10.15 , if $x^{(0)} \in X$ such that $\alpha^{(0)}>1 / h_{0}$, we work with $h \leqslant \Phi /$ opt and after $t=\left\lfloor\log _{1 /(1-h)}\left[h_{0}\left(\alpha^{(0)}-1\right) /\left(1-h_{0}\right)\right]\right\rfloor \leqslant(\mathrm{opt} / \Phi) \cdot \log \left(\alpha_{0} h_{0}\right)$ steps, we obtain $x^{(t)} \in X$ such that $\alpha_{t} \in\left(1,1 / h_{0}\right]$. Otherwise, if $\alpha^{(0)} \in(0,1 / 2)$ we work with $h \leqslant(\Phi /$ opt $) \cdot\left(\alpha_{0} h_{0}\right)^{2}$ and after $t=1 / h$
steps, we obtain $x^{(t)} \in X$ such that $\alpha_{t} \in[1 / 2,1)$. Hence, we can assume that $\alpha_{t} \in\left[1 / 2,1 / h_{0}\right]$ and set $h \leqslant(\Phi /$ opt $) \cdot h_{0}^{2}$. Then, the Lemmas in Section 10.4 and 10.5 are applicable.

Let $E_{1} \stackrel{\text { def }}{=} D\left\|b / \gamma_{A}\right\|_{1}\|c\|_{1}, E_{2} \stackrel{\text { def }}{=} 8 m \rho_{A} \Psi^{(0)}, E_{3} \stackrel{\text { def }}{=} 2 m D^{3} \gamma_{A}\|b\|_{1}$ and $E_{4} \stackrel{\text { def }}{=} 8 m D^{2}\|b\|_{1}$. Consider an arbitrary non-optimal basic feasible solution $g$.

By Lemma 9.10, we have $c^{\mathrm{T}} g \leqslant E_{1}$ and thus both Lemma 10.12 and Lemma 10.15 are applicable with $h, \varepsilon^{\star} \stackrel{\text { def }}{=} \varepsilon / E_{4}$ and any $k \geqslant k_{0} \stackrel{\text { def }}{=} 4 E_{1} /(h \Phi) \cdot \ln \left[\left(E_{2} / \min \left\{1, x_{\min }^{(0)}\right\}\right) \cdot\left(D \gamma_{A} / \varepsilon^{\star}\right)\right]$. Hence, by Lemma 10.15 , the Physarum-inspired dynamics (8.6) guarantees the existence of an index $e \in[m]$ such that $g_{e}>0$ and $x_{e}^{(t+k)}<\varepsilon^{\star} /\left(D \gamma_{A}\right)$. Moreover, by Lemma 10.12 there is a non-negative feasible kernelfree vector $f$ such that $\left\|x^{(t+k)}-f\right\|_{\infty}<\varepsilon^{\star} /\left(D \gamma_{A}\right)$. Thus, for the index $e$ it follows that $g_{e}>0$ and $f_{e}<2 \varepsilon^{\star} / D \gamma_{A}=(\varepsilon / 2) \cdot\left(4 / E_{4} D \gamma_{A}\right)=\varepsilon /\left(2 E_{3}\right)$. Then Lemma 10.14, yields $\left\|f-f^{\star}\right\|_{\infty}<\varepsilon /\left(2 D \gamma_{A}\right)$ and by triangle inequality we have $\left\|x^{(k)}-f^{\star}\right\|_{\infty}<\varepsilon /\left(D \gamma_{A}\right)$.

By construction, $\rho_{A}=\max \left\{D \gamma_{A}, n D^{2}\|A\|_{\infty}\right\} \leqslant n D^{2} \gamma_{A}\|A\|_{\infty}$. Let $E_{2}^{\prime}=8 m n D^{2} \gamma_{A}\|A\|_{\infty} \Psi^{(0)}$ and $E_{5}=E_{2}^{\prime} E_{4} \cdot D \gamma_{A}=8^{2} m^{2} n D^{5} \gamma_{A}^{2}\|A\|_{\infty}\|b\|_{1}$. Further, let $C_{1}=E_{1}$ and $C_{2}=E_{5}$. Then, the statement follows for any $k \geqslant k_{1} \stackrel{\text { def }}{=} 4 C_{1} /(h \Phi) \cdot \ln \left(C_{2} \Psi^{(0)} /\left(\varepsilon \cdot \min \left\{1, x_{\min }^{(0)}\right\}\right)\right)$.

### 10.7 Preconditioning

In this section, we generalize the preconditioning technique developed in $\left[\mathrm{BBD}^{+} 13\right.$, SV 16 b$]$ for flow problems, to the setting of positive linear programs.
Theorem 10.16. Given an integral $L P(A, b, c>0)$, a positive $x^{(0)} \in \mathbb{R}^{m}$ and a parameter $\varepsilon \in(0,1)$. Let $\left([A \mid b], b,\left(c, c^{\prime}\right)\right)$ be an extended $L P$ with $c^{\prime}=2 C_{1}$ and $z^{(0)} \stackrel{\text { def }}{=} 1+D_{S}\|x\|_{\infty}\|A\|_{1}\|b\|_{1} .{ }^{2}$ Then, $\left(x^{(0)} ; z^{(0)}\right)$ is a strongly dominating starting point of the extended problem such that $y^{\mathrm{T}}[A \mid b]\left(x^{(0)}, z^{(0)}\right) \geqslant 1$, for all $y \in Y$. In particular, the Physarum-inspired dynamics (8.6) initialized with $\left(x^{(0)}, z^{(0)}\right)$ and a step size $h \leqslant h_{0}^{2} / C_{3}$, outputs for any $k \geqslant 4 C_{1} \cdot\left(D \gamma_{A}\right)^{2} / h \cdot \ln \left(C_{2} \Upsilon^{(0)} /\left(\varepsilon \cdot \min \left\{1, x_{\min }^{(0)}\right\}\right)\right)$ a vector $\left(x^{(k)}, z^{(k)}\right)>0$ such that $\operatorname{dist}\left(x^{(k)}, X_{\star}\right)<\varepsilon /\left(D \gamma_{A}\right)$ and $z^{(k)}<\varepsilon /\left(D \gamma_{A}\right)$, where $\Upsilon^{(0)} \stackrel{\text { def }}{=} \max \left\{\Psi^{(0)}, z^{(0)}\right\}$.

Theorem 10.16 subsumes [SV16b, Theorem 1.2] for flow problems by giving a tighter asymptotic convergence rate, since for the transshipment problem $A$ is a totally unimodular matrix and satisfies $D=D_{S}=1, \gamma_{A}=1,\|A\|_{\infty}=1$ and $\Phi=1$. We note that the scalar $z^{(0)}$ depends on the scaled determinant $D_{S}$, see Theorem 8.3.

### 10.7.1 Proof of Theorem 10.16

In the extended problem, we concatenate to matrix $A$ a column equal to $b$ such that the resulting constraint matrix becomes $[A \mid b]$. Let $c^{\prime}$ be the cost and let $x^{\prime}$ be the initial capacity of the newly inserted constraint column. We will determine $c^{\prime}$ and $x^{\prime}$ in the course of the discussion. Consider the dual of the max-flow like LP for the extended problem. It has an additional variable $z^{\prime}$ and reads

$$
\min \left\{x^{\mathrm{T}} z+x^{\prime} z^{\prime}: z \geqslant 0 ; z^{\prime} \geqslant 0 ; z \geqslant A^{\mathrm{T}} y ; z^{\prime} \geqslant b^{\mathrm{T}} y ; b^{\mathrm{T}} y=1\right\} .
$$

In any optimal solution, $z^{\prime}=b^{\mathrm{T}} y=1$ and hence the dual is equivalent to

$$
\begin{equation*}
\min \left\{x^{\mathrm{T}} z+x^{\prime}: z \geqslant 0 ; z \geqslant A^{\mathrm{T}} y ; b^{\mathrm{T}} y=1\right\} \tag{10.15}
\end{equation*}
$$

The strongly dominating set of the extended problem is therefore equal to

$$
\begin{equation*}
X=\left\{\binom{x}{x^{\prime}} \in \mathbb{R}_{>0}^{m+1}: y^{\mathrm{T}}[A \mid b]\binom{x}{x^{\prime}}>0 \text { for all } y \in Y\right\} \tag{10.16}
\end{equation*}
$$

The defining condition translates into $x^{\prime}>-y^{\mathrm{T}} A x$ for all $y \in Y$. We summarize the discussion in the following Lemma.
Lemma 10.17. Given a positive $x \in \mathbb{R}^{m}$, let $\rho \stackrel{\text { def }}{=}\|b\|_{1} D_{S}$ and $x^{\prime} \stackrel{\text { def }}{=} 1+\rho\|A\|_{1}\|x\|_{\infty}$, where $\|A\|_{1} \stackrel{\text { def }}{=}$ $\sum_{i, j}\left|A_{i, j}\right|$ and $D_{S} \stackrel{\text { def }}{=} \max \left\{\left|\operatorname{det}\left(A^{\prime}\right)\right|: A^{\prime}\right.$ is a square sub-matrix of $\left.A\right\}$. Then, $\left(x ; x^{\prime}\right)$ is a strongly dominating starting point of the extended problem such that $y^{\mathrm{T}}[A \mid b]\left(x ; x^{\prime}\right)=y^{\mathrm{T}} A x+x^{\prime} \geqslant 1$, for all $y \in Y$.

[^16]Proof. We show first that $\max _{y \in Y}\|y\|_{\infty} \leqslant \rho$ implies the statement. Let $y \in Y$ be arbitrary. Since $\left|y^{\mathrm{T}} A x\right| \leqslant\|A\|_{1}\|x\|_{\infty}\|y\|_{\infty}$, we have $\max _{y \in Y}\left|y^{\mathrm{T}} A x\right| \leqslant \rho\|A\|_{1}\|x\|_{\infty}=x^{\prime}-1$ and hence $y^{\mathrm{T}}[A \mid b]\left(x ; x^{\prime}\right) \geqslant 1$.

It remains to show that $\max _{y \in Y}\|y\|_{\infty} \leqslant \rho$. The constraint polyhedron of the dual (10.15) is given in matrix notation as

$$
P^{(\mathrm{ext})} \stackrel{\text { def }}{=}\left\{\binom{z}{y} \in \mathbb{R}^{m+n}:\left[\begin{array}{cc}
I_{m \times m} & -A^{\mathrm{T}} \\
0_{m}^{\mathrm{T}} & b^{\mathrm{T}} \\
I_{m \times m} & 0_{m \times n}
\end{array}\right]\binom{z}{y} \stackrel{\geqslant}{=}\left(\begin{array}{c}
0_{m} \\
1 \\
0_{m}
\end{array}\right)\right\} .
$$

Let us denote the resulting constraint matrix and vector by $M \in \mathbb{R}^{2 m+1 \times m+n}$ and $d \in \mathbb{R}^{2 m+1}$, respectively.
Note that if $b=0$ then the primal LP (10.2) is either unbounded or infeasible. Hence, we consider the non-trivial case when $b \neq 0$. Observe that the polyhedron $P^{(\mathrm{ext})}$ is not empty, since for any $y$ such that $b^{\mathrm{T}} y=1$ there is $z=\max \left\{0, A^{\mathrm{T}} y\right\}$ satisfying $(z ; y) \in P^{(\text {ext })}$. Further, $P^{(\text {ext })}$ does not contain a line (see Section 10.3) and thus $P^{(\text {ext })}$ has at least one extreme point $p^{\prime} \in P^{(\mathrm{ext})}$. As the dual LP (10.3) has a bounded value (the target function is lower bounded by 0 ) and an extreme point exists ( $p^{\prime} \in P^{(\text {ext })}$ ), the optimum is attained at an extreme point $p \in P^{(\mathrm{ext})}$. Moreover, as every extreme point is a basic feasible solution and matrix $M$ has linearly independent columns ( $A$ has full row rank), it follows that $p$ has $m+n$ tight linearly independent constraints.

Let $M_{B(p)} \in \mathbb{R}^{m+n \times m+n}$ be the basis submatrix of $M$ satisfying $M_{B(p)} p=d_{B(p)}$. Since $A, b$ are integral and $M_{B(p)}$ is invertible, using Laplace expansion we have $1 \leqslant\left|\operatorname{det}\left(M_{B(p)}\right)\right| \leqslant\|b\|_{1} D_{S}=\rho$. Let $Q_{i}$ denotes the matrix formed by replacing the $i$-th column of $M_{B(p)}$ by the column vector $d_{B(p)}$. Then, by Cramer's rule, it follows that $\left|y_{i}\right|=\left|\operatorname{det}\left(Q_{i}\right) / \operatorname{det}\left(M_{B(p)}\right)\right| \leqslant\left|\operatorname{det}\left(M_{B(p)}\right)\right| \leqslant \rho$, for all $i \in[n]$.

It remains to fix the cost of the new column. Using Lemma 9.10, opt $\leqslant c^{\mathrm{T}} x^{(k)} \leqslant C_{1}$ for every $k \in \mathbb{N}$, and thus we set $c^{\prime} \stackrel{\text { def }}{=} 2 C_{1}$.

### 10.8 A Simple Lower Bound

Building on [SV16b, Lemma B.1], we give a lower bound on the number of steps required for computing an $\varepsilon$-approximation to the optimum shortest path. In particular, we show that for the Physarum-inspired dynamics (8.6) to compute a point $x^{(k)}$ such that $\operatorname{dist}\left(x^{(k)}, X_{\star}\right)<\varepsilon$, the required number of steps $k$ has to grow linearly in opt $/(h \Phi)$ and $\ln (1 / \varepsilon)$.

Theorem 10.18. Let $(A, b, c)$ be a positive LP instance such that $A=\left[\begin{array}{ll}11\end{array}\right], b=1$ and $c=(\mathrm{opt} \text {, opt }+\Phi)^{\mathrm{T}}$, where opt $>0$ and $\Phi>0$. Then, for any $\varepsilon \in(0,1)$ the discrete directed Physarum-inspired dynamics (8.6) initialized with $x^{(0)}=(1 / 2,1 / 2)$ and any step size $h \in(0,1 / 2]$, requires at least $k=(1 / 2 h) \cdot \max \{\mathrm{opt} / \Phi, 1\}$. $\ln (2 / \varepsilon)$ steps to guarantee $x_{1}^{(k)} \geqslant 1-\varepsilon, x_{2}^{(k)} \leqslant \varepsilon$. Moreover, if $\varepsilon \leqslant \Phi /(2 \mathrm{opt})$ then $c^{\mathrm{T}} x^{(k)} \geqslant(1+\varepsilon)$ opt as long as $k \leqslant(1 / 2 h) \cdot \max \{\mathrm{opt} / \Phi, 1\} \cdot \ln (2 \Phi /(\varepsilon \cdot \mathrm{opt}))$.

Proof. Let $c_{1}=$ opt and $c_{2}=\gamma$ opt, where $\gamma=1+\Phi /$ opt. We first derive closed-form expressions for $x_{1}^{(k)}, x_{2}^{(k)}$, and $x_{1}^{(k)}+x_{2}^{(k)}$. Let $s^{(k)}=\gamma x_{1}^{(k)}+x_{2}^{(k)}$. For any $k \in \mathbb{N}$, we have $q_{1}^{(k)}+q_{2}^{(k)}=1$ and $q_{1}^{(k)} / q_{2}^{(k)}=\left(x_{1}^{(k)} / c_{1}\right) /\left(x_{2}^{(k)} / c_{2}\right)=\gamma x_{1}^{(k)} / x_{2}^{(k)}$. Therefore, $q_{1}^{(k)}=\gamma x_{1}^{(k)} / s^{(k)}$ and $q_{2}^{(k)}=x_{2}^{(k)} / s^{(k)}$, and hence

$$
\begin{equation*}
x_{1}^{(k)}=\left(1+h\left(-1+\gamma / s^{(k-1)}\right)\right) x_{1}^{(k-1)} \quad \text { and } \quad x_{2}^{(k)}=\left(1+h\left(-1+1 / s^{(k-1)}\right)\right) x_{2}^{(k-1)} . \tag{10.17}
\end{equation*}
$$

Further, $x_{1}^{(k)}+x_{2}^{(k)}=(1-h)\left(x_{1}^{(k-1)}+x_{2}^{(k-1)}\right)+h$, and thus by induction $x_{1}^{(k)}+x_{2}^{(k)}=1$ for all $k \in \mathbb{N}$.
Therefore, $s^{(k)} \leqslant \gamma$ for all $k \in \mathbb{N}$ and hence $x_{1}^{(k)} \geqslant x_{1}^{(k-1)}$, i.e. the sequence $\left\{x_{1}^{(k)}\right\}_{k \in \mathbb{N}}$ is increasing and the sequence $\left\{x_{2}^{(k)}\right\}_{k \in \mathbb{N}}$ is decreasing. Moreover, since $h\left(-1+1 / s^{(k-1)}\right) \geqslant h(1-\gamma) / \gamma=-h \Phi /($ opt $+\Phi)$ and using the inequality $1-z \geqslant e^{-2 z}$ for every $z \in[0,1 / 2]$, it follows by (10.17) and induction on $k$ that

$$
x_{2}^{(k)} \geqslant\left(1-\frac{h \Phi}{\mathrm{opt}+\Phi}\right)^{k} x_{2}^{(0)} \geqslant \frac{1}{2} \exp \left\{-k \cdot \frac{2 h \Phi}{\mathrm{opt}+\Phi}\right\}
$$

Thus, $x_{2}^{(k)} \geqslant \varepsilon$ whenever $k \leqslant(1 / 2 h) \cdot(\mathrm{opt} / \Phi+1) \cdot \ln (2 / \varepsilon)$. This proves the first claim.
For the second claim, observe that $c^{\mathrm{T}} x^{(k)}=\mathrm{opt} \cdot x_{1}^{(k)}+\gamma \mathrm{opt} \cdot x_{2}^{(k)}=\mathrm{opt} \cdot\left(1+(\gamma-1) x_{2}^{(k)}\right)$. This is greater than $(1+\varepsilon)$ opt iff $x_{2}^{(k)} \geqslant \varepsilon \cdot$ opt $/ \Phi$. Thus, $c^{\mathrm{T}} x^{(k)} \geqslant(1+\varepsilon)$ opt as long as $k \leqslant(1 / 2 h) \cdot(\mathrm{opt} / \Phi+$ $1) \cdot \ln (2 /(\varepsilon \cdot \mathrm{opt} / \Phi))$.

## Chapter 11

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[^0]:    ${ }^{1}$ The success probability can be further amplified to $1-\delta$ for any $\delta>0$ by running $O(\log (1 / \delta))$ independent trials of the preceding algorithm.
    ${ }^{2}$ Depending on the specific storage scheme additional running time factors may be necessary. For instance, if we store $A$ by adjacency arrays, then enumerating all non-zero entries in time $O\left(\|A\|_{0}\right)$ is straightforward, while accessing any entry $A_{i, j}$ can be performed in time $O(\log n)$ by a binary search over the adjacency array for row $i$, so in this case the stated running time has to be multiplied by a factor $O(\log n)$.
    ${ }^{3}$ An event happens with high probability (w.h.p.) if it has probability at least $1-1 / n^{c}$ for some $c>0$.

[^1]:    ${ }^{1}$ In Section 2.3, we establish an efficient sampling procedure, see Algorithm 2, that further reduces the total overhead for guessing the sizes $\left|C_{y}\right|$ of an optimal clustering to $\left(2^{k} / \varepsilon\right)^{2^{O(k)}} \cdot(\log n)^{(\log \log n)^{0.1}}$.

[^2]:    ${ }^{2}$ The success probability can be further amplified to $1-\delta$ for any $\delta>0$ by running $O(\log (1 / \delta))$ independent trials of the preceding algorithm.
    ${ }^{3}$ ETH postulates that 3 -SAT is not in time $2^{\circ(m)}$. Here we only need the weaker hypothesis that 3 -SAT is not in time $2^{m^{o(1)}}$.

[^3]:    ${ }^{4}$ Given a submatrix $M$ of $A$, and $t$ columns sampled u.a.r. from $M$, we denote by $\widetilde{C}_{y}$ the resulting multiset of column indices with respect to the original matrix $A$.

[^4]:    ${ }^{5}$ Using similar arguments, for any $\alpha \in[0,1]$ the recurrence $T(a, b) \leqslant(1+\alpha)^{b} \cdot X+Y \cdot T(a-1, b)+T(a, b-1)$ is upper bounded by $X \cdot(2 Y)^{a} \cdot\left(b^{1-\alpha}+2\right)^{a} \cdot(1+\alpha)^{b}$. In particular, this implies that $T(a, b) \leqslant X \cdot(2 Y)^{a} \cdot \min \left\{(b+2)^{a}, 2^{a+b}\right\}$ and thus $T\left(2^{k}, \log n\right) \leqslant(2 / \varepsilon)^{2^{O(k)}} \cdot m n \cdot \min \left\{(\log n)^{2^{k}}, n\right\}$.

[^5]:    ${ }^{6}$ An event happens with high probability (w.h.p.) if it has probability at least $1-1 / n^{c}$ for some $c>0$.

[^6]:    ${ }^{1}$ The Eigendecomposition theorem guarantees that all eigenvectors are orthonormal.
    ${ }^{2}$ Given a symmetric matrix $M$ and a number $k$, the Power method approximates the top $k$ eigenvectors of $M$ corresponding to the largest $k$ eigenvalues. Since we seek a good approximation of the bottom $k$ eigenvectors of $\mathcal{L}_{G}=I-\mathcal{A}$, associated with the smallest $k$ eigenvalues, we initialize the Power method with $M=I+\mathcal{A}$.

[^7]:    ${ }^{3}$ SVD abbreviates Singular Value Decomposition, see [Woo14].
    ${ }^{4}$ Note that $\lambda_{k} / 2 \leqslant \widehat{\rho}(k)$, see (5.8). Thus, the assumption implies $\lambda_{k} / 2 \leqslant \widehat{\rho}(k) \leqslant \delta \lambda_{k+1} /\left(2 \cdot 10^{5} \cdot k^{3}\right)$, i.e., there is a substantial gap between the $(k+1)$-th and the $k$-th eigenvalue.

[^8]:    ${ }^{5}$ A graph $G$ has $k$ connected components iff $\lambda_{k}=0$. For any clique $K_{n}$, we have $\lambda_{1}=0$ and $\lambda_{2}=\cdots=\lambda_{n}=1$. Further, when $G$ consists of $k$ cliques $K_{n / k}$ disconnected from each other, then $\lambda_{1}=\cdots=\lambda_{k}=0$ and $\lambda_{k+1}=\cdots=\lambda_{n}=1$.

[^9]:    ${ }^{1}$ We denote by SPSD the class of symmetric positive semi-definite matrices.

[^10]:    ${ }^{1}$ The kernel of a matrix $A$ consists of all solutions to the system $A x=0$.
    ${ }^{2}$ We conjecture that this also holds for the indices $e$ with $c_{e}=0$.

[^11]:    ${ }^{3}$ Lyapunov functions are a main tool for proving convergence of dynamical systems. It is a function $L(t)$ mapping the state $x(t)$ of the system to a non-negative real such that $\dot{L} \leqslant 0$ and $\dot{L}=0$ iff $\dot{x}=0$. It is an "art" to find a Lyapunov function for a concrete dynamical system.

[^12]:    ${ }^{4}$ Using Lemma 9.10, the dependence on $D_{S}$ can be improved to a scale-independent determinant $D$, defined in (8.8). For further details, we refer the reader to Section 10.2.
    ${ }^{5}$ We postpone the definition of strongly dominating capacity vector to Section 10.3. Every scaled feasible solution is strongly dominating. In the shortest path problem, a capacity vector $x$ is strongly dominating if every source-sink cut $(S, \bar{S})$ has positive directed capacity, i.e., $\sum_{a \in E(S, \bar{S})} x_{a}-\sum_{a \in E(\bar{S}, S)} x_{a}>0$.
    ${ }^{6}$ Note that $\left(\gamma_{A}\right)^{n-1} D \leqslant D_{S} \leqslant\left(\gamma_{A}\right)^{n} D$, and thus $D$ yields an exponential improvement over $D_{S}$, whenever $\gamma_{A} \geqslant 2$.

[^13]:    ${ }^{1}$ For the undirected shortest path problem, we drop the equation corresponding to the sink. Then $b$ becomes the negative indicator vector corresponding to the source node. Note that $n$ is one less than the number of nodes of the graph. The basic feasible solutions are the simple undirected source-sink paths. A circulatory solution contains a cycle on which there is flow.

[^14]:    ${ }^{2}$ In the undirected shortest path problem, the $d$ 's are the incidence vectors of the undirected source-sink cuts. Let $S$ be any set of vertices containing $s_{0}$ but not $s_{1}$, and let $1^{S}$ be its associated indicator vector. The cut corresponding to $S$ contains the edges having exactly one endpoint in $S$. Its indicator vector is $d^{S}=\left|A^{\mathrm{T}} \mathbf{1}^{S}\right|$. Then $d_{e}^{S}=1$ iff $|S \cap\{u, v\}|=1$, where $e=(u, v)$ or $e=(v, u)$, and $d_{e}^{S}=0$ otherwise. For a vector $x \geqslant 0,\left(d^{S}\right)^{\mathrm{T}} x$ is the capacity of the source-sink cut ( $S, V \backslash S$ ). In this setting, $C_{\star}$ is the value of a minimum cut.

[^15]:    ${ }^{1}$ In the shortest path problem (recall that $b=e_{1}-e_{n}$ ) the set $Y$ consists of all $y \in\{-1,+1\}^{n}$ such that $y_{1}=$ $1=-y_{n}$, i.e., $y$ encodes a cut with $S=\left\{i: y_{i}=-1\right\}$ and $\bar{S}=\left\{i: y_{i}=+1\right\}$. The condition $y^{\mathrm{T}} A x>0$ translates into $\sum_{a \in E(S, \bar{S})} x_{a}-\sum_{a \in E(\bar{S}, S)} x_{a}>0$, i.e., every source-sink cut must have positive directed capacity.

[^16]:    ${ }^{2}$ We denote by $\|A\|_{1} \stackrel{\text { def }}{=} \sum_{i, j}\left|A_{i, j}\right|$, i.e. we interpret matrix $A$ as a vector and apply to it the standard $\ell_{1}$ norm.

