A Direct Estimation of High Dimensional Stationary Vector Autoregressions

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Abstract

The vector autoregressive (VAR) model is a powerful tool in learning complex time series and has been exploited in many fields. The VAR model poses some unique challenges to researchers: On one hand, the dimensionality, introduced by incorporating multiple numbers of time series and adding the order of the vector autoregression, is usually much higher than the time series length; On the other hand, the temporal dependence structure naturally present in the VAR model gives rise to extra difficulties in data analysis. The regular way in cracking the VAR model is via "least squares" and usually involves adding different penalty terms (e.g., ridge or lasso penalty) in handling high dimensionality. In this manuscript, we propose an alternative way in estimating the VAR model. The main idea is, via exploiting the temporal dependence structure, formulating the estimating problem to a linear program. There is instant advantage of the proposed approach over the lassotype estimators: The estimation equation can be decomposed to multiple sub-equations and accordingly can be solved efficiently using parallel computing. Besides that, we also bring new theoretical insights into the VAR model analysis. So far the theoretical results developed in high dimensions (e.g., Song and Bickel, 2011 and Kock and Callot, 2015) are based on stringent assumptions that are not transparent. Our results, on the other hand, show that the spectral norms of the transition matrices play an important role in estimation accuracy and build estimation and prediction consistency accordingly. Moreover, we provide some experiments on both synthetic and real-world equity data. We show that there are empirical advantages of our method over the lasso-type estimators in parameter estimation and forecasting.

Keywords: transition matrix, multivariate time series, vector autoregressive model, double asymptotic framework, linear program

1. Introduction

The vector autoregressive (VAR) model plays a fundamental role in analyzing multivariate time series data and has many applications in numerous academic fields. The VAR model is heavily used in finance (Tsay, 2005), econometrics (Sims, 1980), and brain imaging data

analysis (Valdés-Sosa et al., 2005). For example, in understanding the brain connectivity network, multiple resting-state functional magnetic resonance imaging (rs-fMRI) data are obtained by consecutively scanning the same subject for approximately a hundred times or more. This naturally produces a high dimensional dependent data and a common strategy in handling such data is via building a vector autoregressive model (see Qiu et al., and the references therein).

This manuscript considers estimating the VAR model. Our focus is on the stationary vector autoregression with the order (or called lag) p and Gaussian noises. More specifically, let random vectors X_1, \ldots, X_T be from a stochastic process $(X_t)_{t=-\infty}^{\infty}$. Each X_t is a d-dimensional random vector and satisfies that

$$X_t = \sum_{k=1}^p A_k^{\mathrm{T}} X_{t-k} + Z_t, \quad Z_t \sim N_d(0, \Psi),$$

where A_1, \ldots, A_p are called the transition matrices and $(Z_t)_{t=-\infty}^{\infty}$ are independent multivariate Gaussian noises. Via assuming $\det(I_d - \sum_{k=1}^p A_k^{\mathrm{T}} z^k) \neq 0$ for all $z \in \mathcal{C}$ with modulus not greater than one, we then have the process is stationary (check, for example, Section 2.1 in Lütkepohl, 2005) and $X_t \sim N_d(0, \Sigma)$ for some covariance matrix Σ depending on $\{A_k, k=1, \ldots, p\}$ and Ψ .

There are in general three main targets in analyzing an VAR model. One is to estimate the transition matrices A_1, \ldots, A_p . These transition matrices reveal the temporal dependence in the data sequence and estimating them builds a fundamental first step in forecasting. Moreover, the zero and nonzero entries in the transition matrices directly incorporate the Granger non-causalities and causalities with regard to the stochastic sequence (see, for example, Corollary 2.2.1 in Lütkepohl, 2005). Another one of interest is the error covariance Ψ , which reveals the contemporaneous interactions among d time series. Finally, by merely treating the temporal dependence as another measure of the data dependence (in parallel to the mixing conditions, Bradley, 2005), it is also of interest to estimate the covariance matrix Σ .

This manuscript focuses on estimating the transition matrices A_1, \ldots, A_p , while noting that the techniques developed here can also be exploited to estimate the covariance matrix Σ and the noise covariance Ψ . We first review the methods developed so far in transition matrix estimation. Let $A = (A_1^T, \ldots, A_p^T)^T \in \mathbb{R}^{dp \times d}$ be the combination of the transition matrices. Given X_1, \ldots, X_T , the perhaps most classic method in estimating A is least squares minimization (Hamilton, 1994)

$$\widehat{A}^{\text{LSE}} = \underset{M \in \mathbb{R}^{dp \times d}}{\operatorname{argmin}} \| \widetilde{Y} - M^{\mathsf{T}} \widetilde{X} \|_{\mathsf{F}}^{2}, \tag{1}$$

where $\|\cdot\|_{\mathsf{F}}$ is the matrix Frobenius norm, $\widetilde{Y} = (X_{p+1}, \dots, X_T) \in \mathbb{R}^{d \times (T-p)}$, and $\widetilde{X} = \{(X_p^{\mathsf{T}}, \dots, X_1^{\mathsf{T}})^{\mathsf{T}}, \dots, (X_{T-1}^{\mathsf{T}}, \dots, X_{T-p}^{\mathsf{T}})^{\mathsf{T}}\} \in \mathbb{R}^{(dp) \times (T-p)}$. However, a fatal problem in (1) is that the product of the order of the autoregression p and the number of time series d is frequently larger than the time series length T. Therefore, the model has to be constrained to enforce identifiability. A common strategy is to add sparsity on the transition matrices so that the number of nonzero entries is less than T. Built on this assumption, there has been a large literature discussing adding different penalty terms to (1) for regularizing

the estimator: From the ridge-penalty to the lasso-penalty and more non-concave penalty terms. In the following we list the major efforts. Hamilton (1994) discussed the use of the ridge-penalty $||M||_{\mathsf{F}}^2$ in estimating the transition matrices. Hsu et al. (2008) proposed to add the L_1 -penalty in estimating the transition matrices, inducing a sparse output. Several extensions to transition matrix estimation in the VAR model include: Wang et al. (2007) exploited the L_1 -penalty in simultaneously estimating the regression coefficients and determining the number of lags in a linear regression model with autoregressive errors. In detecting causality, Haufe et al. (2008) transferred the problem to estimating transition matrices in an VAR model and advocated using a group-lasso penalty for inducing joint sparsity among a whole block of coefficients. In studying the graphical Granger causality problem, Shojaie and Michailidis (2010) exploited the VAR model and proposed to estimate the coefficients using a truncated weighted L_1 -penalty. Song and Bickel (2011) exploited the L_1 penalty in a complicated VAR model and aimed to select the variables and lags simultaneously.

The theoretical properties of the L_1 -regularized estimator have been analyzed in Bento et al. (2010), Nardi and Rinaldo (2011), Song and Bickel (2011), and Kock and Callot (2015) under the assumption that the matrix A is sparse, i.e., the number of nonzero entries in A is much less than the dimension of parameters pd^2 . Nardi and Rinaldo (2011) provided both subset and parameter estimation consistency results under a relatively low dimensional settings with $d = o(n^{1/2})$. Bento et al. (2010) studied the problem of estimating supports sets of the transition matrices in the high dimensional settings and proposed an "irrepresentable condition" similar as what is proposed in the linear regression model (Zou, 2006; Zhao and Yu, 2006; Meinshausen and Bühlmann, 2006; Wainwright, 2009). It is for the L_1 regularized estimator to attain the support set selection consistency. In parallel, Song and Bickel (2011) and Kock and Callot (2015) studied the parameter estimation and support set selection consistency of the L_1 -regularized estimator in high dimensions.

In this paper, we propose a new approach to estimate the transition matrix A. Different from the line of lasso-based estimation procedures, which are built on penalizing the least square term, we exploit the linear programming technique and the proposed method is very fast to solve via parallel computing. Moreover, we do not need A to be exactly sparse and allow it to be only "weakly sparse". The main idea is to estimate A using the relationship between A and the marginal and lag 1 autocovariance matrices (such a relationship is referred to as the Yule-Walker equation). We thus formulate the estimation procedure to a linear program, while adding the $\|\cdot\|_{\text{max}}$ (element-wise supremum norm) for model identifiability. Here we note that the proposed procedure can be considered as a generalization of the Dantzig selector (Candes and Tao, 2007) to the linear regression model with multivariate response. Indeed, our proposed method can also be exploited in conducting multivariate regression (Breiman and Friedman, 1997).

The proposed method enjoys several advantages compared to the existing ones: (i) Computationally, our method can be formulated into d linear programs and can be solved in parallel. Similar ideas have been used in learning high dimensional linear regression (Candes and Tao, 2007; Bickel et al., 2009) and graphical models (Yuan, 2010; Cai et al., 2011). (ii) In the model-level, our method allows A to be only weakly sparse. (iii) Theoretically, so far the analysis on lasso-type estimators (Song and Bickel, 2011; Kock and Callot, 2015) depends on certain regularity conditions, restricted eigenvalue conditions on the design

matrix for example, which are not transparent and do not explicitly reveal the role of temporal dependence in it. In contrast, we provide explicit nonasymptotic analysis, and our analysis highlights the spectral norm $||A||_2$ in estimation accuracy, which is inspired by some recent developments (Loh and Wainwright, 2012). Moreover, for exact sign recovery, our analysis does not need the "irrepresentable condition" which is usually required in the analysis of lasso-type estimators (Bento et al., 2010).

The major theoretical results are briefly stated as follows. We adopt a double asymptotic framework where d is allowed to increase with T. We call a matrix s-sparse if there are at most s nonzero elements on each of its column. Under mild conditions, we provide the explicit rates of convergence of our estimator \widehat{A} based on the assumption that A is s-sparse (Cai et al., 2011). In particular, for lag 1 time series, we show that

$$\|\widehat{A} - A\|_1 = O_P \left\{ \frac{s\|A\|_1}{1 - \|A\|_2} \left(\frac{\log d}{T} \right)^{1/2} \right\}, \quad \|\widehat{A} - A\|_{\max} = O_P \left\{ \frac{\|A\|_1}{1 - \|A\|_2} \left(\frac{\log d}{T} \right)^{1/2} \right\},$$

where $\|\cdot\|_{\text{max}}$ and $\|\cdot\|_q$ represent the matrix elementwise absolute maximum norm (L_{max} norm) and induced L_q norm (detailed definitions will be provided in §2). Using the L_{max} norm consistency result, we further provide the sign recovery consistency of the proposed method. This result is of self interest and sheds light to detecting Granger causality. We also provide the prediction consistency results based on the L_1 consistency result and show that element-wise error in prediction can be controlled. Here for simplicity we only provide the results when A is exactly sparse and defer the presentation of the results for weakly sparse matrix to Section 4.

The rest of the paper is organized as follows. In Section 2, we briefly review the vector autoregressive model. In Section 3, we introduce the proposed method for estimating the transition matrices of the vector autoregressive model. In Section 4, we provide the main theoretical results. In Section 5, we apply the new method to both synthetic and real equity data for illustrating its effectiveness. More discussions are provided in the last section. Detailed technical proofs are provided in the appendix¹.

2. Background

In this section, we briefly review the vector autoregressive model. Let $M = (M_{jk}) \in \mathbb{R}^{d \times d}$ and $v = (v_1, ..., v_d)^T \in \mathbb{R}^d$ be a matrix and an vector of interest. We denote v_I to be the subvector of v whose entries are indexed by a set $I \subset \{1, ..., d\}$. We also denote $M_{I,J}$ to be the submatrix of M whose rows are indexed by I and columns are indexed by I. We denote $M_{I,*}$ to be the submatrix of M whose rows are indexed by I, $M_{*,J}$ to be the submatrix of M whose columns are indexed by I. For $0 < q < \infty$, we define the L_0 , L_q , and L_∞ vector (pseudo-)norms to be

$$||v||_0 := \sum_{j=1}^d I(v_j \neq 0), \quad ||v||_q := \left(\sum_{j=1}^d |v_j|^q\right)^{1/q}, \quad \text{and} \quad ||v||_\infty := \max_{1 \leq j \leq d} |v_j|,$$

^{1.} Some of the results in this paper were first stated without proof in a conference version (Han and Liu, 2013).

where $I(\cdot)$ is the indicator function. Letting M be a matrix, we denote the matrix L_q , L_{max} , and Frobenius norms to be

$$||M||_q := \max_{||v||_q = 1} ||Mv||_q, \quad ||M||_{\max} := \max_{jk} |M_{jk}|, \text{ and } ||M||_{\mathsf{F}} := \left(\sum_{j,k} |M_{jk}|^2\right)^{1/2}.$$

We denote $\mathbf{1}_d = (1, \dots, 1)^{\mathrm{T}} \in \mathbb{R}^d$. Let $\sigma_1(M) \geq \dots \geq \sigma_d(M)$ be the singular values of M.

Let $p \geq 1$ be an integer. A lag p vector autoregressive process can be elaborated as follows: Let $(X_t)_{t=-\infty}^{\infty}$ be a stationary sequence of random vectors in \mathbb{R}^d with mean 0 and covariance matrix Σ . We say that $(X_t)_{t=-\infty}^{\infty}$ follow a lag p vector autoregressive model if and only if they satisfy

$$X_{t} = \sum_{k=1}^{p} A_{k}^{\mathrm{T}} X_{t-k} + Z_{t} \quad (t \in \mathbb{Z}).$$
 (2)

Here A_1, \ldots, A_p are called transition matrices. We denote $A = (A_1^{\mathsf{T}}, \ldots, A_p^{\mathsf{T}})^{\mathsf{T}}$ to be the combination of the transition matrices. We assume that Z_t are independently and identically generated from a Gaussian distribution $N_d(0, \Psi)$. Moreover, Z_t and $(X_s)_{s < t}$ are independent for any $t \in \mathbb{Z}$. We pose an additional assumption that $\det(I_d - \sum_{k=1}^p A_k^{\mathsf{T}} z^k) \neq 0$ for all $z \in \mathcal{C}$ with modulus not greater than one. This guarantees that the sequence is stationary and we have, for any $t \in \mathbb{Z}$, X_t follows a Gaussian distribution $N_d(0, \Sigma)$,

We denote $\Sigma_i(\cdot)$ to be an operator on the process $(X_t)_{t=-\infty}^{\infty}$. In particular, we define $\Sigma_i\{(X_t)\} = \operatorname{Cov}(X_0, X_i)$. It is easy to see that $\Sigma_0\{(X_t)\} = \Sigma$. If the lag of the vector autoregressive model is 1 (i.e., $X_t = A_1^{\mathrm{T}} X_{t-1} + Z_t$, for any $t \in \mathbb{Z}$), by simple calculation we have the so called "Yule-Walker Equation"

$$\Sigma_i\{(X_t)\} = \Sigma_0\{(X_t)\}(A_1)^i, \tag{3}$$

which further implies that

$$A_1 = [\Sigma_0\{(X_t)\}]^{-1} \cdot \Sigma_1\{(X_t)\}.$$

The results for lag 1 vector autoregressive model can be extended to the lag p vector autoregressive model by appropriately redefining the random vectors. In detail, the autoregressive model with lag p shown in (2) can be reformulated as an autoregressive model with lag 1

$$\widetilde{X}_t = \widetilde{A}^{\mathrm{T}} \widetilde{X}_{t-1} + \widetilde{Z}_t, \tag{4}$$

where

$$\widetilde{X}_{t} = \begin{pmatrix} X_{t+p-1} \\ X_{t+p-2} \\ \vdots \\ X_{t} \end{pmatrix}, \quad \widetilde{A} = \begin{pmatrix} A_{1} & I_{d} & 0 & \dots & 0 \\ \vdots & \ddots & \dots & \ddots & \vdots \\ A_{p-1} & 0 & 0 & \dots & I_{d} \\ A_{p} & 0 & 0 & \dots & 0 \end{pmatrix}, \quad \widetilde{Z}_{t} = \begin{pmatrix} Z_{t+p-1} \\ 0 \\ \vdots \\ 0 \end{pmatrix}.$$
(5)

Here $I_d \in \mathbb{R}^{d \times d}$ is the identity matrix, $\widetilde{X}_t \sim N_{dp}(0, \widetilde{\Sigma})$ for $t = 1, \ldots, T$, and $\widetilde{Z}_t \sim N_{dp}(0, \widetilde{\Psi})$ with $\widetilde{\Sigma} = \text{Cov}(\widetilde{X}_t)$ and $\widetilde{\Psi} = \text{Cov}(\widetilde{Z}_t)$. Therefore, we also have

$$\widetilde{A} = [\Sigma_0\{(\widetilde{X}_t)\}]^{-1} \cdot \Sigma_1\{(\widetilde{X}_t)\}.$$
(6)

This is similar to the relationship for the lag 1 vector autoregressive model.

3. Methods and Algorithms

We provide a new formulation to estimate A_1, \ldots, A_p for the vector autoregressive model. Let X_1, \ldots, X_T be from a lag p vector autoregressive process $(X_t)_{t=-\infty}^{\infty}$ and we denote $\widetilde{X}_t = (X_{t+p-1}^T, \ldots, X_t^T)^T$ for $t = 1, \ldots, T-p+1$. We denote S and S_1 to be the marginal and lag 1 sample covariance matrices of $(\widetilde{X}_t)_{t=1}^{T-p+1}$

$$S := \frac{1}{T - p + 1} \sum_{t=1}^{T - p + 1} \widetilde{X}_t \widetilde{X}_t^{\mathrm{T}}, \quad S_1 := \frac{1}{T - p} \sum_{t=1}^{T - p} \widetilde{X}_t \widetilde{X}_{t+1}^{\mathrm{T}}.$$
 (7)

Using the connection between \widetilde{A} and $\Sigma_0\{(\widetilde{X}_t)\}, \Sigma_1\{(\widetilde{X}_t)\}$ shown in (6), we know that a good estimator $\widetilde{\Omega}$ of \widetilde{A} shall satisfy that

$$\|\Sigma_0\{(\widetilde{X}_t)\}\check{\Omega} - \Sigma_1\{(\widetilde{X}_t)\}\| \tag{8}$$

is small enough with regard to a certain matrix norm $\|\cdot\|$. Moreover, using the fact that $A = (A_1^T, \ldots, A_p^T)^T = \widetilde{A}_{*,J}$, where $J = \{1, \ldots, d\}$, by (8) we have that a good estimate \check{A} of A shall satisfy

$$\|\Sigma_0\{(\widetilde{X}_t)\}\check{A} - [\Sigma_1\{(\widetilde{X}_t)\}]_{*,J}\|$$
(9)

is small enough.

Motivated by (9), we estimate A_1, \ldots, A_p via replacing $\Sigma_0\{(\widetilde{X}_t) \text{ and } [\Sigma_1\{(\widetilde{X}_t)\}]_{*,J}$ with their empirical versions. For formulating the estimation equation to a linear program, we use the L_{max} norm. Accordingly, we end in solving the following convex optimization program

$$\widehat{\Omega} = \underset{M \in \mathbb{R}^{dp \times d}}{\operatorname{argmin}} \sum_{jk} |M_{jk}|, \text{ subject to } ||SM - (S_1)_{*,J}||_{\max} \le \lambda_0,$$
(10)

where $\lambda_0 > 0$ is a tuning parameter. In (10), the constraint part aims to find an estimate that approximates the true parameter well, and combined with the minimization part, aims to induce certain sparsity. Let $\widehat{\Omega}_{*,j} = \widehat{\beta}_j$, it is easy to see that (10) can be decomposed to many subproblems and each $\widehat{\beta}_j$ can be solved by

$$\widehat{\beta}_j = \underset{v \in \mathbb{R}^{dp}}{\operatorname{argmin}} \|v\|_1, \quad \text{subject to} \quad \|Sv - (S_1)_{*,j}\|_{\infty} \le \lambda_0. \tag{11}$$

Accordingly, compared to the lasso-type procedures, the proposed method can be solved in parallel and therefore is computationally more efficient.

Once $\widehat{\Omega}$ is obtained, the estimator of the transition matrix A_k can then be written as

$$\widehat{A}_k = \widehat{\Omega}_{J_k,*},\tag{12}$$

where we denote $J_k = \{j : d(k-1) + 1 \le j \le dk\}.$

We now show that the optimization in (11) can be formulated into a linear program. Recall that any real number a takes the decomposition $a=a^+-a^-$, where $a^+=a\cdot I (a\geq 0)$ and $a^-=-a\cdot I (a<0)$. For any vector $v=(v_1,\ldots,v_d)^{\rm T}\in\mathbb{R}^d$, let $v^+=(v_1^+,\ldots,v_d^+)^{\rm T}$ and $v^-=(v_1^-,\ldots,v_d^-)^{\rm T}$. We denote $v\geq 0$ if $v_1,\ldots,v_d\geq 0$ and v<0 if $v_1,\ldots,v_d<0$, $v_1\geq v_2$

if $v_1 - v_2 \ge 0$, and $v_1 < v_2$ if $v_1 - v_2 < 0$. Letting $v = (v_1, \dots, v_d)^T$, the problem in (11) can be further relaxed to the following problem

$$\widehat{\beta}_{j} = \underset{v^{+}, v^{-}}{\operatorname{argmin}} \mathbf{1}_{d}^{\mathsf{T}}(v^{+} + v^{-}),$$
subject to $||Sv^{+} - Sv^{-} - (S_{1})_{*,j}||_{\infty} \le \lambda_{0}, \ v^{+} \ge 0, v^{-} \ge 0.$ (13)

To minimize $\mathbf{1}_d^{\mathrm{T}}(v^+ + v^-)$, v^+ or v^- can not be both nonzero. Therefore, the solution to (13) is exactly the solution to (11). The optimization in (13) can be written as

$$\widehat{\beta}_{j} = \operatorname*{argmin}_{v^{+}, v^{-}} \mathbf{1}_{d}^{\mathsf{T}}(v^{+} + v^{-}),$$
 subject to $Sv^{+} - Sv^{-} - (S_{1})_{*,j} \leq \lambda_{0} \mathbf{1}_{d},$
$$-Sv^{+} + Sv^{-} + (S_{1})_{*,j} \leq \lambda_{0} \mathbf{1}_{d},$$

$$v^{+} > 0, v^{-} > 0.$$

This is equivalent to

$$\widehat{\beta}_j = \underset{\omega}{\operatorname{argmin}} \mathbf{1}_{2d}^{\mathrm{T}} \omega, \quad \text{subject to} \quad \theta + W \omega \ge 0, \quad \omega \ge 0, \tag{14}$$

where

$$\omega = \begin{pmatrix} v^+ \\ v^- \end{pmatrix}, \quad \theta = \begin{bmatrix} (S_1)_{*,j} + \lambda_0 \mathbf{1}_d \\ -(S_1)_{*,j} + \lambda_0 \mathbf{1}_d \end{bmatrix}, \quad W = \begin{pmatrix} -S & S \\ S & -S \end{pmatrix}.$$

The optimization (14) is a linear program. We can solve it using the simplex algorithm (Murty, 1983).

4. Theoretical Properties

In this section, under the double asymptotic framework, we provide the nonasymptotic rates of convergence in parameter estimation under the matrix L_1 and L_{max} norms.

We first present the rates of convergence of the estimator $\widehat{\Omega}$ in (10) under the vector autoregressive model with lag 1. This result allows us to sharply characterize the impact of the temporal dependence of the time series on the obtained rate of convergence. In particular, we show that the rate of convergence is closely related to the L_1 and L_2 norms of the transition matrix A_1 , where $||A_1||_2$ is the key part in characterizing the impact of temporal dependence on estimation accuracy. Secondly, we present the sign recovery consistency result of our estimator. Compared to the lasso-type estimators, our result does not require the irrepresentable condition. These results are combined together to show that we have the prediction consistency, i.e., the term $||A_1X_T - \widehat{A}_1X_T||$ goes to zero with regard to certain norms $||\cdot||$. The application to lag p > 1 case is left for future studies.

We start with some additional notation. Let $M_d \in \mathbb{R}$ be a quantity which may scale with the time series length and dimension (T, d). We define the set of square matrices in $\mathbb{R}^{d \times d}$, denoted by $\mathcal{M}(q, s, M_d)$, as

$$\mathcal{M}(q, s, M_d) := \Big\{ M \in \mathbb{R}^{d \times d} : \max_{1 \le j \le d} \sum_{i=1}^{d} |M_{ij}|^q \le s, ||M||_1 \le M_d \Big\}.$$

For q = 0, the class $\mathcal{M}(0, s, M_d)$ contains all the s-sparse matrices with bounded L_1 norms. There are two general remarks about the model $\mathcal{M}(q, s, M_d)$: (i) $\mathcal{M}(q, s, M_d)$ can be considered as the matrix version of the vector "weakly sparse set" explored in Raskutti et al. (2011) and Vu and Lei (2012). Such a way to define the weakly sparse set of matrices is also investigated in Cai et al. (2011). (ii) For the exactly sparse matrix set, $\mathcal{M}(0, s, M_d)$, the sparsity level s here represents the largest number of nonzero entries in each column

the sparsity level s here represents the largest number of nonzero entries in each column of the matrix. In contrast, the sparsity level s' exploited in Kock and Callot (2015) is the total number of nonzero entries in the matrix. We must have $s' \geq s$ and regularly $s' \gg s$ (means $s/s' \to 0$).

The next theorem presents the L_1 and L_{max} rates of convergence of our estimator under the vector autoregressive model with lag 1.

Theorem 1 Suppose that $(X_t)_{t=1}^T$ are from a lag 1 vector autoregressive process $(X_t)_{t=-\infty}^\infty$ as described in (2). We assume the transition matrix $A_1 \in \mathcal{M}(q, s, M_d)$ for some $0 \le q < 1$. Let \widehat{A}_1 be the optimum to (10) with the tuning parameter

$$\lambda_0 = \frac{32\|\Sigma\|_2 \max_j(\Sigma_{jj})}{\min_j(\Sigma_{jj})(1 - \|A\|_2)} (2M_d + 3) \left(\frac{\log d}{T}\right)^{1/2}.$$

For $T \geq 6 \log d + 1$ and $d \geq 8$, we have, with probability no smaller than $1 - 14d^{-1}$,

$$\|\widehat{A}_{1} - A_{1}\|_{1} \leq 4s \left\{ \frac{32\|\Sigma^{-1}\|_{1} \max_{j}(\Sigma_{jj})\|\Sigma\|_{2}}{\min_{j}(\Sigma_{jj})(1 - \|A_{1}\|_{2})} (2M_{d} + 3) \left(\frac{\log d}{T}\right)^{1/2} \right\}^{1-q}.$$
 (15)

Moreover, with probability no smaller than $1 - 14d^{-1}$,

$$\|\widehat{A}_1 - A_1\|_{\max} \le \frac{64\|\Sigma^{-1}\|_1 \max_j(\Sigma_{jj})\|\Sigma\|_2}{\min_j(\Sigma_{jj})(1 - \|A_1\|_2)} (2M_d + 3) \left(\frac{\log d}{T}\right)^{1/2}.$$
 (16)

In the above results, Σ is the marginal covariance matrix of X_t .

It can be observed that, similar to the lasso and Dantzig selector (Candes and Tao, 2007; Bickel et al., 2009), the tuning parameter λ_0 here depends on the variance term Σ . In practice, same as most preceded developments (see, for example, Song and Bickel, 2011), we can use a data-driven way to select the tuning parameter. In this manuscript we explore using cross-validation to choose λ_0 with the best prediction accuracy. In Section 5 we will show that the procedure of selecting the tuning parameter via cross-validation gives reasonable results.

Here A_1 is assumed to be at least weakly sparse and belong to the set $\mathcal{M}(q, s, M_d)$. This is merely for the purpose of model identifiability. Otherwise, we will have multiple global optima in the optimization problem.

The obtained rates of convergence in Theorem 1 depend on both Σ and A_1 with $||A_1||_2$ characterizing the temporal dependence. In particular, the estimation error is related to the spectral norm of the transition matrix A_1 . Intuitively, this is because $||A_1||_2$ characterizes the data dependence of X_1, \ldots, X_T , and accordingly intrinsically characterizes how much information there is in the data. If $||A_1||_2$ is larger, then there is less information we can

exploit in estimating A_1 . Technically, $||A_1||_2$ determines the rate of convergence of S and S_1 to their population counterparts. We refer to the proofs of Lemmas 1 and 2 for details.

In the following, we list two examples to provide more insights about the results in Theorem 1.

Example 1 We consider the case where Σ is a strictly diagonal dominant (SDD) matrix (Horn and Johnson, 1990) with the property

$$\delta_i := |\Sigma_{ii}| - \sum_{j \neq i} |\Sigma_{ij}| \ge 0, \quad (i = 1, \dots, d).$$

This corresponds to the cases where the d entries in any X_t with $t \in \{1, ..., T\}$ are weakly dependent. In this setting, Ahlberg and Nilson (1963) showed that

$$\|\Sigma^{-1}\|_{1} = \|\Sigma^{-1}\|_{\infty} \le \left\{ \min_{i} \left(|\Sigma_{ii}| - \sum_{j \neq i} |\Sigma_{ij}| \right) \right\}^{-1} = \max_{i} (\delta_{i}^{-1}). \tag{17}$$

Moreover, by algebra, we have

$$\|\Sigma\|_{2} \le \|\Sigma\|_{1} = \max_{i} \left(|\Sigma_{ii}| + \sum_{j \ne i} |\Sigma_{ij}| \right) \le 2 \max_{i} (|\Sigma_{ii}|).$$
 (18)

Equations (17) and (18) suggest that, when $\max_i(\Sigma_{ii})$ is upper bounded, and both $\min_i(\Sigma_{ii})$ and δ_i are lower bounded by a fixed constant, we have both $\|\Sigma^{-1}\|_1$ and $\|\Sigma\|_2$ are upper bounded, and the obtained rates of convergence in (15) and (16) can be simplified as

$$\|\widehat{A}_1 - A_1\|_1 = O_P \left[s \left\{ \frac{M_d}{1 - \|A_1\|_2} \left(\frac{\log d}{T} \right)^{1/2} \right\}^{1 - q} \right],$$

$$\|\widehat{A}_1 - A_1\|_{\max} = O_P \left\{ \frac{M_d}{1 - \|A_1\|_2} \left(\frac{\log d}{T} \right)^{1/2} \right\}.$$

Example 2 We can generalize the "entry-wise weakly dependent" structure in Example 1 to a "block-wise weakly dependent" structure. More specifically, we consider the case where $\Sigma = (\Sigma_{jk}^b)$ with blocks $\Sigma_{jk}^b \in \mathbb{R}^{d_j \times d_k}$ $(1 \leq j \leq K)$ is a strictly block diagonal dominant (SBDD) matrix with the property

$$\delta_i^b = \|(\Sigma_{ii}^b)^{-1}\|_{\infty}^{-1} - \sum_{i \neq i} \|\Sigma_{ij}^b\|_{\infty} > 0 \quad (i = 1, \dots, K).$$

In this case, Varah (1975) showed that

$$\|\Sigma^{-1}\|_1 = \|\Sigma^{-1}\|_{\infty} \le \left\{ \min_{i} \left(\|(\Sigma_{ii}^b)^{-1}\|_{\infty}^{-1} - \sum_{i \neq i} \|\Sigma_{ij}^b\|_{\infty} \right) \right\}^{-1} = \max\{(\delta_i^b)^{-1}\}.$$

Moreover, we have

$$\|\Sigma\|_{2} \le \|\Sigma\|_{1} \le \max_{i} (\|(\Sigma_{ii}^{b})^{-1}\|_{\infty}^{-1} + \|\Sigma_{ii}^{b}\|_{\infty}).$$

Accordingly, generally ($\|(\Sigma_{ii}^b)^{-1}\|_{\infty}^{-1} + \|\Sigma_{ii}^b\|_{\infty}$) is in the scale of $\max_i(d_i) \ll d$, and when δ_i^b are lower bounded and the condition number of Σ is upper bounded, we have the obtained rates of convergence can be simplified as

$$\|\widehat{A}_1 - A_1\|_1 = O_P \left[s \left\{ \frac{M_d \cdot \max_i(d_i)}{1 - \|A_1\|_2} \left(\frac{\log d}{T} \right)^{1/2} \right\}^{1-q} \right],$$

$$\|\widehat{A}_1 - A_1\|_{\max} = O_P \left\{ \frac{M_d \cdot \max_i(d_i)}{1 - \|A_1\|_2} \left(\frac{\log d}{T} \right)^{1/2} \right\}.$$

We then continue to the results of feature selection. If we have $A_1 \in \mathcal{M}(0, s, M_d)$, from the element-wise L_{max} norm convergence, a sign recovery result can be obtained. In detail, let \check{A}_1 be a truncated version of \widehat{A}_1 with level γ

$$(\check{A}_1)_{ij} = (\widehat{A}_1)_{ij} I\{|(\widehat{A}_1)_{ij}| \ge \gamma\}.$$
 (19)

The following corollary shows that \check{A}_1 recovers the sign of A_1 with overwhelming probability.

Corollary 1 Suppose that the conditions in Theorem 1 hold and $A_1 \in \mathcal{M}(0, s, M_d)$. If we choose the truncation level

$$\gamma = \frac{64\|\Sigma^{-1}\|_1 \max_j(\Sigma_{jj})\|\Sigma\|_2}{\min_j(\Sigma_{jj})(1 - \|A_1\|_2)} (2M_d + 3) \left(\frac{\log d}{T}\right)^{1/2}$$

in (19) and with the assumption that

$$\min_{\{(j,k):(A_1)_{jk}\neq 0\}} |(A_1)_{jk}| \ge 2\gamma,$$

we have, with probability no smaller than $1 - 14d^{-1}$, $\operatorname{sign}(A_1) = \operatorname{sign}(\check{A}_1)$. Here for any matrix M, $\operatorname{sign}(M)$ is a matrix with each element representing the sign of the corresponding entry in M.

Here we note that Corollary 1 sheds lights to detecting Granger causality. For any two processes $\{y_t\}$ and $\{z_t\}$, Granger (1969) defined the causal relationship in principle as follows: Provided that we know everything in the universe, $\{y_t\}$ is said to cause $\{z_t\}$ in Granger's sense if removing the information about $\{y_s\}_{s \leq t}$ from the whole knowledge base built by time t will increase the prediction error about z_t . It is known that the noncausalities are determined by the transition matrices in the stable VAR process (Lütkepohl, 2005). Therefore, detecting the nonzero entries of A_1 consistently means that we can estimate the Granger-causality network consistently.

We then turn to evaluate the prediction performance of the proposed method. Given a new data point X_{T+1} in the time point T+1, based on $(X_t)_{t=1}^T$, the next corollary quantifies the distance between X_{T+1} and \widehat{A}_1X_T in terms of L_{∞} norm.

Corollary 2 Suppose that the conditions in Theorem 1 hold and let

$$\Psi_{\max} := \max_{i} (\Psi_{ii})$$
 and $\Sigma_{\max} := \max_{i} (\Sigma_{ii}).$

Then for the new data point X_{T+1} at time point T+1 and any constant $\alpha > 0$, with probability greater than

$$1 - 2(d^{\alpha/2 - 1}\sqrt{\pi/2 \cdot \alpha \log d})^{-1} - 14d^{-1},$$

we have

$$||X_{T+1} - \widehat{A}_1^{\mathrm{T}} X_T||_{\infty} \le (\Psi_{\max} \cdot \alpha \log d)^{1/2} + 4s \left\{ \frac{32||\Sigma^{-1}||_1 \max_j(\Sigma_{jj})||\Sigma||_2}{\min_j(\Sigma_{jj})(1 - ||A_1||_2)} (2M_d + 3) \left(\frac{\log d}{T}\right)^{1/2} \right\}^{1-q} \cdot (\Sigma_{\max} \cdot \alpha \log d)^{1/2}, \quad (20)$$

where \widehat{A}_1 is calculated based on $(X_t)_{t=1}^T$.

Here we note that the first term in the right-hand side of Equation (20), $(\Psi_{\text{max}} \cdot \alpha \log d)^{1/2}$, is present due to the diverges of the new data point from its mean caused by an unpredictable noise perturb term $Z_{T+1} \sim N_d(0, \Psi)$. This term is unable to be canceled out even if we have almost infinite data points. The second term in the right-hand side of Equation (20) depends on the estimation accuracy of \widehat{A}_1 to A_1 and will converge to zero under certain conditions. In other words, the term

$$||A_1^{\mathrm{T}} X_T - \widehat{A}_1^{\mathrm{T}} X_T||_{\infty} \to 0,$$

converges to zero in probability as $n, d \to \infty$.

Although A_1 is in general asymmetric, there exist cases such that a symmetric transition matrix is more of interest. It is known that the off-diagonal entries in the transition matrix represent the influence of one state on the others and such influence might be symmetric or not. Weiner et al. (2012) provided several examples where a symmetric transition matrix is more appropriate for modeling the data.

If we can further suppose that the transition matrix A_1 is symmetric, we can use this information and obtain a new estimator \bar{A}_1 as

$$(\bar{A}_1)_{jk} = (\bar{A}_1)_{kj} := (\hat{A}_1)_{jk} I(|(\hat{A}_1)_{jk}| \le |(\hat{A}_1)_{kj}|) + (\hat{A}_1)_{kj} I(|(\hat{A}_1)_{kj}| \le |(\hat{A}_1)_{jk}|).$$

In other word, we always pick the entry with smaller magnitudes. Then using Theorem 1, we have $\|\bar{A}_1 - A_1\|_1$ and $\|\bar{A}_1 - A_1\|_{\infty}$ can be upper bounded by the same number presented in the right-hand side of (15). In this case, because both A_1 and \bar{A}_1 are symmetric, we have $\|\bar{A}_1 - A_1\|_2 \leq \|\bar{A}_1 - A_1\|_1 = \|\bar{A}_1 - A_1\|_{\infty}$. We then proceed to quantify the prediction accuracy under L_2 norm in the next corollary.

Corollary 3 Suppose that the conditions in Theorem 1 hold and A_1 is a symmetric matrix. Then for the new data point X_{T+1} at time point T+1, with probability greater than $1-18d^{-1}$, we have

$$||X_{T+1} - \bar{A}_1^{\mathrm{T}} X_T||_2 \le \sqrt{2||\Psi||_2 \log d} + \sqrt{\operatorname{tr}(\Psi)} + 4s \left\{ \frac{32||\Sigma^{-1}||_1 \max_j(\Sigma_{jj})||\Sigma||_2}{\min_j(\Sigma_{jj})(1 - ||A_1||_2)} (2M_d + 3) \left(\frac{\log d}{T}\right)^{1/2} \right\}^{1-q} \cdot \left\{ \sqrt{2||\Sigma||_2 \log d} + \sqrt{\operatorname{tr}(\Sigma)} \right\}.$$
(21)

Based on Corollary 3, we have, similar as what is discussed in Corollary 2, the term $||A_1^T X_T - \widehat{A}_1^T X_T||_2$ will vanish when the second term in the left-hand side of (21) can converge to zero.

5. Experiments

We conduct numerical experiments on both synthetic and real data to illustrate the effectiveness of our proposed method compared to the competing ones, as well as obtain more insights on the performance of the proposed method. In the following we consider the three competing methods:

- (i) The least square estimation using a ridge-penalty (The method in Hamilton, 1994, by adding a ridge-penalty $||M||_F^2$ to the least squares loss function in Equation 1).
- (ii) The least square estimation using an L_1 penalty (The method in Hsu et al., 2008, by adding an L_1 penalty $\sum_{ij} |M_{ij}|$ to Equation 1).
- (iii) Our method (The estimator described in Equation 10).

Here we consider including the procedure discussed in Hamilton (1994) because it is a commonly explored baseline and shows how bad the classic procedure can be when the dimension is high. We only consider the competing procedure proposed in Hsu et al. (2008) because this is the only method that is specifically designed for the same simple VAR as what we study. We do not consider other aforementioned procedures (e.g., Haufe et al., 2008; Shojaie and Michailidis, 2010) because they are designed for more specific models with more assumptions. We use the R package "glmnet" (Friedman et al., 2010) for implementing the lasso method in Hsu et al. (2008), and the simplex algorithm for implementing ours.

5.1 Cross-Validation Procedure

We start with an introduction to how to conduct cross-validation for choosing the lag p and the tuning parameter λ in the algorithm outlined in Section 3.

For the time series $(X_t)_{t=-\infty}^T$ and a specific time point t_0 of interest, if both p and λ are assumed to be unknown, the proposed cross-validation procedure is as follows.

- 1. We set all possible choices of (p, λ) to be a grid. We set n_1 and n_2 to be two numbers (representing the length of training data and the number of replicates).
- 2. For each X_t among $X_{t_0-1}, \ldots, X_{t_0-n_2}$, the estimates $\widehat{A}_1^t(p,\lambda), \ldots, \widehat{A}_p^t(p,\lambda)$ are calculated based on the training data $X_{t-1}, \ldots, X_{t-n_1}$ and any choice of (p,λ) . We set the prediction error at time t, denoted as $\operatorname{Err}_t(p,\lambda)$, to be $\operatorname{Err}_t(p,\lambda) := \|X_t \sum_{k=1}^p \widehat{A}_k^t(p,\lambda)^{\mathrm{T}} X_{t-k}\|_2$.
- 3. We take an average over the prediction errors and denote

$$\overline{\mathrm{Err}}(p,\lambda) := \frac{1}{n_2} \sum_{t=t_0-n_2}^{t_0-1} \mathrm{Err}_t(p,\lambda)$$

4. We choose the (p, λ) over the grid such that $\overline{\mathrm{Err}}(p, \lambda)$ is minimized.

In case when p is predetermined, the above procedure can be easily modified to focus only on selecting λ with p to be the determined value.

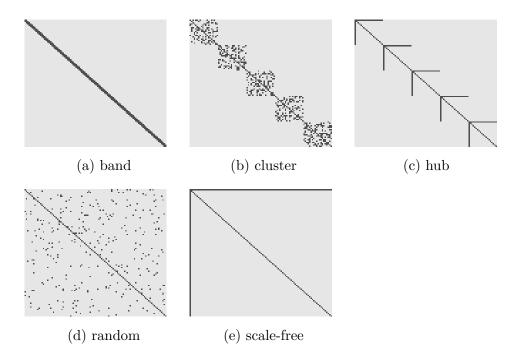


Figure 1: Five different transition matrix patterns used in the experiments. Here gray points represent the zero entries and black points represent nonzero entries.

5.2 Synthetic Data Analysis

In this subsection, we compare the performance of our method with the ridge and lasso methods using synthetic data under multiple settings. We also study the impact of transition matrices' spectral norms on estimation accuracy, and how the computation time and memory usage of all methods scale with the number of lags.

5.2.1 Performance Comparison: Lag p=1

This section focuses on vector autoregressive model described in (2) with lag one. We compare our method to the competing ones on several synthetic data sets. We consider the settings where the time series length T varies from 50 to 100 and the dimension d varies from 50 to 200.

We create the transition matrix A_1 according to five different patterns: band, cluster, hub, random, and scale-free. Typical realizations of these patterns are illustrated in Figure 1 and are generated using the "flare" package in R (Li et al., 2015). In those plots, the gray points represent the zero entries and the black points represent the nonzero entries. We then rescale A_1 such that we have $||A_1||_2 = 0.5$. Once A_1 is obtained, we generate Σ using two models. First is the simple setting with Σ to be diagonal

$$\Sigma = 2||A_1||_2 I_d. \tag{22}$$

		rio	ridge method			sso methor	od	our method		
d	T	L_{F}	L_2	L_1	L_{F}	L_2	L_1	L_{F}	L_2	L_1
50	100	2.71	0.52	2.47	2.34	0.50	1.54	2.08	0.49	0.58
		(0.028)	(0.023)	(0.103)	(0.064)	(0.029)	(0.161)	(0.045)	(0.006)	(0.039)
100	50	4.21	0.64	3.54	5.52	0.75	3.13	3.26	0.52	1.03
		(0.026)	(0.024)	(0.136)	(0.075)	(0.024)	(0.211)	(0.052)	(0.017)	(0.321)
200	100	7.28	0.76	6.26	6.36	0.64	2.77	4.26	0.50	0.69
		(0.031)	(0.018)	(0.132)	(0.057)	(0.015)	(0.112)	(0.045)	(0.003)	(0.035)

Table 1: Comparison of estimation performance of three methods with diagonal covariance matrix over 1,000 replications. The standard deviations are presented in the parentheses. Here L_{F}, L_2 , and L_1 represent the Frobenius, L_2 , and L_1 matrix norms respectively. The pattern of the transition matrix is "band".

The second is the complex setting where Σ is of Toeplitz form

$$\Sigma_{i,i} = 1$$
, $\Sigma_{i,j} = \rho^{|i-j|}$ for some $\rho \in (0,1)$ and $i, j = 1, \dots, d$.

We then calculate the covariance matrix Ψ of the Gaussian noise vector Z_t as $\Psi = \Sigma - A_1^T \Sigma A_1$. With A_1, Σ , and Ψ , we simulate a time series $(X_1, \ldots, X_T)^T \in \mathbb{R}^{T \times d}$ according to the model described in (2).

We construct 1,000 replicates and compare the three methods described above. The averaged estimation errors under different matrix norms are illustrated in Tables 1 to 10. The standard deviations of the estimation errors are provided in the parentheses. The tuning parameters for the three methods are selected using the cross-validation procedure outlined in Section 5.1 with $n_1 = T/2$, $n_2 = T/2$, and the lag p predetermined to be 1.

Tables 1 to 10 show that our method nearly uniformly outperforms the methods in Hsu et al. (2008) and Hamilton (1994) under different norms (Frobenius, L_2 , and L_1 norms). In particular, the improvement over the method in Hsu et al. (2008) tends to be more significant when the dimension d is larger. Our method also has averagely slightly less standard deviations compared to the method in Hsu et al. (2008), but overall the difference is not significant. The method in Hamilton (1994) has worse performance than the other two methods. This verifies that it is not appropriate to handle very high dimensional data.

5.2.2 Synthetic Data: Lag $p \ge 1$

In this section, we further compare the performance of the three competing methods under the settings of possibly multiple lags, with the number of lags known.

In detail, we choose p to be from 1 to 9, the time series length T=100, and the dimension d=50. The transition matrices A_1, \ldots, A_p are created according to "hub" or "scale-free" pattern, and then rescaled such that $||A_i||_2 = 0.1$ for $i=1,\ldots,p$. The error covariance matrix Ψ is set to be identity for simplicity. Under this multiple lags setting, we then calculate the covariance matrix of \widetilde{X}_t , i.e., $\widetilde{\Sigma}$ defined in (5), by solving a discrete Lyapunov

		rio	ridge method			sso meth	od	our method		
d	T	L_{F}	L_2	L_1	L_{F}	L_2	L_1	L_{F}	L_2	L_1
50	100	2.48	0.44	2.40	2.12	0.43	1.56	1.48	0.49	0.69
		(0.034)	(0.024)	(0.110)	(0.055)	(0.032)	(0.119)	(0.020)	(0.011)	(0.026)
100	50	3.74	0.58	3.46	5.24	0.67	3.16	2.27	0.50	0.66
		(0.031)	(0.022)	(0.121)	(0.084)	(0.025)	(0.223)	(0.002)	(0.001)	(0.002)
200	100	6.80	0.72	6.26	5.82	0.55	2.80	3.02	0.49	0.77
		(0.025)	(0.021)	(0.188)	(0.058)	(0.014)	(0.109)	(0.024)	(0.010)	(0.047)

Table 2: Comparison of estimation performance of three methods with diagonal covariance matrix over 1,000 replications. The standard deviations are presented in the parentheses. Here L_{F}, L_2 , and L_1 represent the Frobenius, L_2 , and L_1 matrix norms respectively. The pattern of the transition matrix is "cluster".

		rio	ridge method			sso meth	od	our method		
d	T	L_{F}	L_2	L_1	L_{F}	L_2	L_1	L_{F}	L_2	L_1
50	100	2.41	0.42	2.37	1.96	0.38	1.48	1.16	0.41	1.05
		(0.033)	(0.027)	(0.102)	(0.06)	(0.039)	(0.141)	(0.115)	(0.058)	(0.092)
100	50	3.49	0.55	3.44	5.06	0.63	3.11	1.86	0.50	1.40
		(0.034)	(0.023)	(0.143)	(0.088)	(0.032)	(0.214)	(0.118)	(0.016)	(0.138)
200	100	6.61	0.69	6.24	5.48	0.52	2.75	2.12	0.50	1.26
		(0.035)	(0.017)	(0.133)	(0.062)	(0.019)	(0.147)	(0.046)	(0.006)	(0.031)

Table 3: Comparison of estimation performance of three methods with diagonal covariance matrix over 1,000 replications. The standard deviations are presented in the parentheses. Here L_{F}, L_2 , and L_1 represent the Frobenius, L_2 , and L_1 matrix norms respectively. The pattern of the transition matrix is "hub".

		rio	ridge method			sso methor	od	our method		
d	T	L_{F}	L_2	L_1	L_{F}	L_2	L_1	L_{F}	L_2	L_1
50	100	2.60	0.48	2.45	2.21	0.43	1.53	1.73	0.44	0.73
		(0.031)	(0.027)	(0.102)	(0.061)	(0.030)	(0.143)	(0.051)	(0.026)	(0.034)
100	50	4.10	0.61	3.53	5.44	0.71	3.09	3.07	0.48	1.21
		(0.025)	(0.020)	(0.136)	(0.077)	(0.024)	(0.224)	(0.066)	(0.024)	(0.177)
200	100	7.01	0.74	6.27	6.03	0.58	2.79	3.54	0.44	0.95
		(0.024)	(0.019)	(0.179)	(0.048)	(0.011)	(0.163)	(0.036)	(0.026)	(0.079)

Table 4: Comparison of estimation performance of three methods with diagonal covariance matrix over 1,000 replications. The standard deviations are presented in the parentheses. Here L_{F}, L_2 , and L_1 represent the Frobenius, L_2 , and L_1 matrix norms respectively. The pattern of the transition matrix is "random".

		rio	ridge method			sso metho	od	our method		
d	T	L_{F}	L_2	L_1	L_{F}	L_2	L_1	L_{F}	L_2	L_1
50	100	2.48	0.44	2.40	2.09	0.41	1.51	1.44	0.41	0.98
		(0.032)	(0.025)	(0.098)	(0.059)	(0.033)	(0.154)	(0.075)	(0.052)	(0.108)
100	50	3.60	0.56	3.43	5.14	0.64	3.11	2.16	0.46	1.36
		(0.034)	(0.023)	(0.133)	(0.085)	(0.031)	(0.188)	(0.130)	(0.043)	(0.115)
200	100	6.65	0.70	6.26	5.57	0.51	3.29	2.51	0.42	2.49
		(0.034)	(0.017)	(0.143)	(0.065)	(0.014)	(0.274)	(0.249)	(0.050)	(0.108)

Table 5: Comparison of estimation performance of three methods with diagonal covariance matrix over 1,000 replications. The standard deviations are presented in the parentheses. Here L_{F}, L_2 , and L_1 represent the Frobenius, L_2 , and L_1 matrix norms respectively. The pattern of the transition matrix is "scale-free".

		rio	ridge method			sso meth	od	our method		
d	T	L_{F}	L_2	L_1	L_{F}	L_2	L_1	L_{F}	L_2	L_1
50	100	2.47	0.51	2.25	2.10	0.45	1.32	1.82	0.47	0.57
		(0.031)	(0.033)	(0.101)	(0.066)	(0.035)	(0.131)	(0.084)	(0.014)	(0.044)
100	50	3.98	0.67	3.31	5.22	0.74	2.81	3.15	0.51	1.04
		(0.029)	(0.033)	(0.107)	(0.083)	(0.032)	(0.174)	(0.114)	(0.063)	(0.529)
200	100	6.92	0.79	5.96	5.82	0.61	2.44	3.79	0.48	0.67
		(0.033)	(0.028)	(0.142)	(0.060)	(0.023)	(0.134)	(0.078)	(0.006)	(0.034)

Table 6: Comparison of estimation performance of three methods on data generated with Toeplitz covariance matrix ($\rho = 0.5$), over 1,000 replications. The standard deviations are presented in the parentheses. Here $L_{\rm F}, L_2$, and L_1 represent the Frobenius, L_2 , and L_1 matrix norms respectively. The pattern of the transition matrix is "band".

		rio	ridge method			sso metho	od	our method			
d	T	L_{F}	L_2	L_1	L_{F}	L_2	L_1	L_{F}	L_2	L_1	
50	100	2.32	0.42	2.25	2.01	0.39	1.42	1.46	0.47	0.69	
		(0.041)	(0.029)	(0.114)	(0.066)	(0.030)	(0.124)	(0.027)	(0.019)	(0.037)	
100	50	3.61	0.57	3.33	5.08	0.65	3.01	2.47	0.47	1.02	
		(0.034)	(0.029)	(0.124)	(0.087)	(0.031)	(0.212)	(0.075)	(0.031)	(0.155)	
200	100	6.63	0.70	6.13	5.58	0.54	2.59	2.96	0.48	0.79	
		(0.038)	(0.020)	(0.162)	(0.069)	(0.019)	(0.153)	(0.027)	(0.013)	(0.046)	

Table 7: Comparison of estimation performance of three methods on data generated with Toeplitz covariance matrix ($\rho = 0.5$), over 1,000 replications. The standard deviations are presented in the parentheses. Here $L_{\rm F}, L_2$, and L_1 represent the Frobenius, L_2 , and L_1 matrix norms respectively. The pattern of the transition matrix is "cluster".

		rio	ridge method			sso metho	od	our method		
d	T	L_{F}	L_2	L_1	L_{F}	L_2	L_1	L_{F}	L_2	L_1
50	100	2.27	0.40	2.22	1.85	0.36	1.34	1.16	0.39	1.01
		(0.039)	(0.037)	(0.099)	(0.067)	(0.041)	(0.157)	(0.124)	(0.062)	(0.102)
100	50	3.37	0.54	3.26	4.94	0.61	2.96	1.86	0.50	1.37
		(0.041)	(0.034)	(0.125)	(0.102)	(0.033)	(0.222)	(0.120)	(0.017)	(0.104)
200	100	6.46	0.67	6.19	5.24	0.50	2.54	2.13	0.49	1.24
		(0.042)	(0.024)	(0.168)	(0.071)	(0.025)	(0.162)	(0.107)	(0.023)	(0.042)

Table 8: Comparison of estimation performance of three methods on data generated with Toeplitz covariance matrix ($\rho = 0.5$), over 1,000 replications. The standard deviations are presented in the parentheses. Here L_{F}, L_2 , and L_1 represent the Frobenius, L_2 , and L_1 matrix norms respectively. The pattern of the transition matrix is "hub".

		rio	ridge method			sso methor	od	our method			
d	T	L_{F}	L_2	L_1	L_{F}	L_2	L_1	L_{F}	L_2	L_1	
50	100	2.49	0.45	2.34	2.15	0.41	1.44	1.74	0.44	0.74	
		(0.036)	(0.029)	(0.104)	(0.071)	(0.032)	(0.139)	(0.058)	(0.033)	(0.043)	
100	50	4.02	0.60	3.42	5.34	0.70	2.96	3.07	0.47	1.21	
		(0.029)	(0.024)	(0.123)	(0.092)	(0.028)	(0.207)	(0.085)	(0.027)	(0.192)	
200	100	6.89	0.72	6.13	5.87	0.56	2.65	3.54	0.43	0.97	
		(0.028)	(0.022)	(0.164)	(0.057)	(0.016)	(0.174)	(0.052)	(0.019)	(0.091)	

Table 9: Comparison of estimation performance of three methods on data generated with Toeplitz covariance matrix ($\rho = 0.5$), over 1,000 replications. The standard deviations are presented in the parentheses. Here $L_{\rm F}, L_2$, and L_1 represent the Frobenius, L_2 , and L_1 matrix norms respectively. The pattern of the transition matrix is "random".

		rio	ridge method			sso methor	od	our method		
d	T	L_{F}	L_2	L_1	L_{F}	L_2	L_1	L_{F}	L_2	$\overline{L_1}$
50	100	2.36	0.42	2.27	2.00	0.38	1.36	1.42	0.37	0.89
		(0.036)	(0.033)	(0.094)	(0.064)	(0.033)	(0.136)	(0.068)	(0.056)	(0.108)
100	50	3.49	0.55	3.29	5.03	0.63	2.96	2.21	0.42	1.29
		(0.039)	(0.029)	(0.124)	(0.100)	(0.027)	(0.212)	(0.149)	(0.050)	(0.131)
200	100	6.52	0.67	6.18	5.36	0.49	3.06	2.55	0.39	$\bf 2.44$
		(0.041)	(0.019)	(0.165)	(0.070)	(0.013)	(0.219)	(0.364)	(0.062)	(0.134)

Table 10: Comparison of estimation performance of three methods on data generated with Toeplitz covariance matrix ($\rho = 0.5$), over 1,000 replications. The standard deviations are presented in the parentheses. Here $L_{\rm F}, L_2$, and L_1 represent the Frobenius, L_2 , and L_1 matrix norms respectively. The pattern of the transition matrix is "scale-free".

equation $\widetilde{A}^{\mathsf{T}}\widetilde{\Sigma}\widetilde{A} - \widetilde{\Sigma} + \widetilde{\Psi} = 0$. This is via using the Matlab command "dlyapchol". With $\{A_i\}_{i=1}^p, \widetilde{\Sigma}$, and Ψ determined, we simulate a time series $(X_1, \ldots, X_T)^{\mathsf{T}} \in \mathbb{R}^{T \times d}$ according to the model described in (2) (with lag $p \geq 1$).

The estimation error is calculated by measuring the difference of $(A_1^{\rm T},\ldots,A_p^{\rm T})^{\rm T}$ and $(\widehat{A}_1^{\rm T},\ldots,\widehat{A}_p^{\rm T})^{\rm T}$ with regard to different matrix norms $(L_{\rm F},L_2,$ and L_1 norms). We conduct 1,000 simulations and compare the averaged performance of three competing methods. The calculated averaged estimation errors are illustrated in Tables 11 and 12. The standard deviations of the estimation errors are provided in the parentheses. Here the tuning parameters are selected in the same way as before. Tables 11 and 12 confirms that our method still outperforms the competing two methods.

5.2.3 Synthetic Data: Impact of Transition Matrices' Spectral Norms

In this section we illustrate the effects of the transition matrices' spectral norms on estimation accuracy. To this end, we study the settings in Section 5.2. More specifically, we set lag p=1, the dimension d and the sample size T to be d=50 and T=100. The transition matrix A_1 is created according to different patterns ("band", "cluster", "hub", "scale-free", and "random"), and then rescaled such that $||A_1||_2 = \kappa$, where κ is from 0.05 to 0.9. Covariance matrix Σ is set to be of the form (22), and Ψ is accordingly determined by stationary condition. We select the tuning parameters using the cross-validation procedure as before. The estimation errors are then plotted against κ and shown in Figure 2.

Figure 2 illustrates that the estimation error is an increasing function of the spectral norm $||A_1||_2$. This demonstrates that the spectral norms of the transition matrices play an important role in estimation accuracy and justifies the theorems in Section 4.

	rio	dge meth	od	la	sso metho	bc	0	ur metho	d
p	L_{F}	L_2	L_1	L_{F}	L_2	L_1	L_{F}	L_2	L_1
1	6.93	2.50	7.35	1.83	0.52	1.36	0.25	0.11	0.23
	(0.012)	(0.094)	(0.377)	(0.039)	(0.017)	(0.128)	(0.014)	(0.016)	(0.002)
3	9.13	2.89	15.96	2.52	0.59	2.18	0.45	0.18	0.70
	(0.129)	(0.092)	(0.249)	(0.085)	(0.016)	(0.116)	(0.023)	(0.004)	(0.003)
5	5.57	1.57	11.73	2.75	0.61	3.19	0.58	0.23	1.23
	(0.000)	(0.000)	(0.000)	(0.000)	(0.000)	(0.000)	(0.000)	(0.000)	(0.000)
7	4.27	1.14	10.92	2.90	0.60	3.44	0.72	0.31	1.83
	(0.010)	(0.041)	(0.152)	(0.026)	(0.025)	(0.183)	(0.077)	(0.067)	(0.222)
9	3.59	0.90	10.17	2.98	0.61	4.11	0.70	0.30	2.11
	(0.026)	(0.023)	(0.219)	(0.061)	(0.004)	(0.201)	(0.000)	(0.000)	(0.000)

Table 11: Comparison of estimation performance of three methods over 1,000 replications under multiple lag settings. The standard deviations are presented in the parentheses. Here L_{F}, L_2 , and L_1 represent the Frobenius, L_2 , and L_1 matrix norms respectively. The pattern of the transition matrix is "hub".

	rio	dge meth	od	la	sso metho	od	our method			
p	L_{F}	L_2	L_1	L_{F}	L_2	L_1	L_{F}	L_2	L_1	
1	6.93	2.51	7.39	1.83	0.53	1.35	0.30	0.12	0.24	
	(0.116)	(0.093)	(0.340)	(0.041)	(0.018)	(0.129)	(0.045)	(0.016)	(0.039)	
3	9.14	3.00	15.97	2.53	0.60	2.19	0.46	0.17	0.57	
	(0.133)	(0.099)	(0.219)	(0.090)	(0.020)	(0.094)	(0.058)	(0.007)	(0.083)	
5	5.58	1.57	11.66	2.77	0.60	2.97	0.62	0.23	0.93	
	(0.002)	(0.002)	(0.018)	(0.001)	(0.002)	(0.076)	(0.012)	(0.002)	(0.078)	
7	4.28	1.14	10.97	2.90	0.60	3.34	0.69	0.24	1.29	
	(0.014)	(0.042)	(0.164)	(0.031)	(0.020)	(0.131)	(0.041)	(0.005)	(0.078)	
9	3.62	0.90	10.25	3.01	0.61	3.42	0.87	0.30	1.79	
	(0.024)	(0.023)	(0.267)	(0.058)	(0.003)	(0.112)	(0.078)	(0.012)	(0.198)	

Table 12: Comparison of estimation performance of three methods over 1,000 replications under multiple lag settings. The standard deviations are presented in the parentheses. Here L_{F}, L_2 , and L_1 represent the Frobenius, L_2 , and L_1 matrix norms respectively. The pattern of the transition matrix is "scale-free".

Estimation error O.0 0.5 1.0 1.5 2.0 A cluster Hub A cluster Landom Landom A cluster Landom Landom Landom Landom Landom Landom Landom Landom Landom

Estimation Error v.s Spectral Norm of A₁

Figure 2: Estimation errors of A_1 (in L_1 norm) plotted against spectral norms of A_1 .

Spectral norm of A₁

0.4

0.6

8.0

5.2.4 Computation Time and Memory Usage

0.0

0.2

This section is devoted to show the computation time and memory usage of our method. First, we show the advantage of our method in terms of saving the computation time. A major advantage of our method over the two competing methods is that our method can be easily parallelly computed, and hence has the potential to save the computation time. We illustrate this point with a figure and two tables using the computation time as a function of the number of available cores. All experiments are conducted on a 2816-core Westmere/Ivybridge 2.67/2.5GHz Linux server with 17T memory, a cluster system with batch scheduling.

We first focus on the lag p=1 case. In detail, we set the time series length T=100 and the dimension d=50. The transition matrix A_1 is created according to the pattern "random", and then rescaled such that $||A_1||_2=0.5$. The covariance matrix Σ is generated as in (22), and Ψ is generated by stationary condition. We then solve (11) using parallel computing based on 1 to 50 cores.

Figure 3 shows the computation time. It illustrates that, in terms of saving the computation time, under this specific setting, we have: (i) Our method outperforms the ridge method even if we do not parallelly compute it; (ii) When there are no less than 8 cores, our method outperforms the lasso method. Here the ridge method is very slow because it involves calculating the inverse of a large matrix.

To further study the advantage of parallel computing when the number of lags, p, grows, we provide another experiment focusing on models with varying lags. More specifically, we

Time Usage v.s # of Available Cores

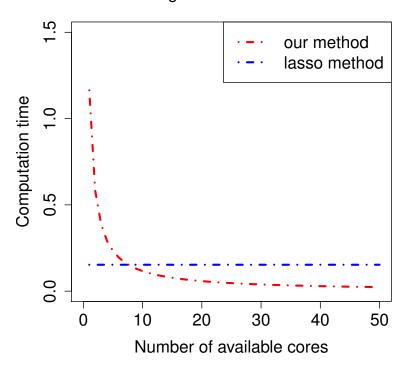


Figure 3: Computation time v.s number of available cores. The computation time for the ridge and lasso methods are 5.843s and 0.153s, which do not change with number of available cores. The computation time here is the averaged elapsed time (in seconds) of 100 replicates of a single experiment.

	lasso method	01	our method (with # of available cores)							
p	N/A	1	5	10	20	30	40	50		
1	0.260	1.277	0.261	0.132	0.067	0.048	0.033	0.029		
2	0.664	2.732	0.553	0.280	0.141	0.098	0.073	0.059		
3	1.034	8.945	1.792	0.897	0.455	0.299	0.230	0.181		
4	1.538	18.278	3.695	1.844	0.920	0.620	0.466	0.366		
5	1.946	35.609	7.130	3.890	1.781	1.189	0.870	0.719		

Table 13: A comparison of computation time with increasing number of lags p: lasso method v.s our method. The computation time for the lasso method does not change with number of available cores. The computation time here is the averaged elapsed time (in seconds) of 100 replicates of a single experiment.

set the time series length T=100 and the dimension d=50. The transition matrices A_1, \ldots, A_p are created according to the pattern "random", and then rescaled such that $||A_i||_2 = 0.1$ for $i=1,\ldots,p$. The error covariance matrix Ψ and the covariance matrix $\widetilde{\Sigma}$ are generated in the same way as in Section 5.2.2. With $\{A_i\}_{i=1}^p, \widetilde{\Sigma}$, and Ψ determined, we simulate a time series $(X_1,\ldots,X_T)^{\mathrm{T}} \in \mathbb{R}^{T\times d}$ according to the model described in (2). We then solve (11) using parallel computing based on 1 to 50 cores.

Table 13 lists the averaged elapsed time of 100 replicates of one single experiment. Here for each replication, the parameters $(A_1, \ldots, A_p, \Psi, \widetilde{\Sigma})$ in the experiment are regenerated. It illustrates that, in terms of saving the computation time, under this specific setting, we have: (i) When there is only one core, the lasso method outperforms our method. But when there are no less than 20 cores, our method outperforms the lasso method for all lags p=1,2,3,4,5; (ii) As p grows, the advantage of parallel computing will be less significant (The ratio of computation time between our method at the maximum number of available cores and the lasso method tends to increase). We also observe from Table 13 that: (iii) The computation time of our method is approximately increasing quadratically with regard to the lag p, while the computation time of the lasso method is approximately increasing linearly with regard to the lag p.

Similarly, to study the advantage of parallel computing when the dimension d grows, we provide an experiment focusing on models with varying dimensions. We consider the settings where the length T = 100, the lag p = 1, and the dimension d varies from 10 to 200. The transition matrix A_1 is created according to the pattern "random", and then rescaled such that $||A_1||_2 = 0.5$. The covariance matrix Σ is generated as in (22), and Ψ is generated by stationary condition. We then solve (11) using parallel computing based on 1 to 200 cores.

Similar to Table 13, Table 14 lists the computation time. It illustrates that, in terms of saving the computation time, under this specific setting, we have: (i) When there is only one core, the lasso method outperforms our method. But when using the maximum number of cores (i.e., d cores), our method outperforms the lasso method for all lags p = 1, 2, 3, 4, 5; (ii)

	lasso method	our method (with $\#$ of available cores))	
d	N/A	1	5	10	20	50	100	200
10	0.022	0.074	0.015	0.008	N/A	N/A	N/A	N/A
20	0.048	0.265	0.055	0.027	0.014	N/A	N/A	N/A
50	0.153	1.164	0.234	0.120	0.061	0.027	N/A	N/A
100	0.468	6.354	1.281	0.649	0.318	0.131	0.067	N/A
200	2.320	21.503	4.304	2.157	1.111	0.448	0.219	0.108

Table 14: A comparison of computation time with increasing dimension d: lasso method v.s our method. The computation time for the lasso method does not change with number of available cores. The computation time here is the averaged elapsed time (in seconds) of 100 replicates of a single experiment.

As d grows, the advantage of parallel computing will be more significant (The ratio between our method at the maximum number of available cores and the lasso method decreases).

Tables 13 and 14 illustrate that, when p or d grows, the advantage of parallel computing becomes less or more significant respectively. Such results are reasonable because (3.4) can be decomposed to at most d subproblems in a columnwise way, and solved in parallel. As the dimension d grows, the maximum number of decomposed subproblems accordingly grows, and hence the gain in parallel computing will be more significant. In comparison, as p grows (while d is fixed), the maximum number of subproblems does not grow, and hence the advantage of parallel computing is less significant.

Secondly, we show the memory usage of our method. By converting the time series from VAR(1) to VAR(p) or increasing the dimension d, the memory usage increases. For investigating the memory usage, we conduct an empirical study. Specifically, first, we choose the lag p to be $1,2,\ldots,9$, the time series length T=100, and the dimension d=50. Transition matrices A_1,\ldots,A_p are created according to the "random" pattern, and then rescaled such that $\|A_i\|_2=0.1$ for $i=1,\ldots,p$. Ψ is set as I_d for simplicity. With $\{A_i\}_{i=1}^p$ and Ψ , we simulate a time series $(X_1,\ldots,X_T)^T\in\mathbb{R}^{T\times d}$ according to (2) with lag $p\geq 1$. The first two rows in Table 15 reports the averaged memory usage of 100 replicates of one single experiment in megabytes (Mb). Here for each replication, the parameters in the experiment are regenerated.

Secondly, we choose the lag p=1, the time series length T=100, the dimension d=50, and the transition matrix A_1 to be created according to the "random" pattern, and then rescaled such that $||A_1||_2=0.1$. Ψ is set as I_d for simplicity. With A_1 and Ψ , we simulate a time series $(X_1,\ldots,X_T)^T\in\mathbb{R}^{T\times d}$ according to (2). The second two rows in Table 15 reports the memory usage.

Table 15 shows that, under this setting, the memory usage is approximately increasing linearly with regard to p; On the other hand, the memory usage is approximately increasing quadratically with regard to d, and this pattern becomes clearer when d is larger.

Lag of model (p)	1	2	3	4	5	6	7	8	9
Mem. Use (Mb)	5.566	8.724	11.862	14.999	18.135	21.272	24.406	27.540	30.673
Dimension (d)	10	20	30	40	50	75	100	150	200
Mem. Use (Mb)	1.649	2.235	3.083	4.194	5.566	10.150	16.390	33.754	57.678

Table 15: Memory usage v.s lag of model and dimension: The result shown below is the averaged memory usage (in Mb) of 100 replicates of one single experiment, with the lag p changing from 1 to 9 or dimension changing from 10 to 200. The pattern of the transition matrices $\{A_i\}_{i=1}^p$ is "random".

5.3 Real Data

We further compare the three methods on the equity data collected from Yahoo! Finance. The task is to predict the stock prices. We collect the daily closing prices for 91 stocks that are consistently in the S&P 100 index between January 1, 2003 and January 1, 2008. This gives us altogether 1,258 data points, each of which corresponds to the vector of closing prices on a trading day.

We first provide comparison on averaged prediction errors for using different lag p on this data set. Let $E = (E_{t,j}) \in \mathbb{R}^{1258 \times 91}$ with $E_{t,j}$ denoting the closing price of the stock j on day t. We screen out all the stocks with low marginal standard deviations and only keep 50 stocks which vary the most. We center the data so that the marginal mean of each time series is zero. The resulting data matrix is denoted by $\bar{E} \in \mathbb{R}^{1258 \times 50}$. We apply the three methods on \bar{E} with different lag p changing from 1 to 9. To evaluate the performance of the three methods, for $t = 1248, \ldots, 1257$, we select the data set $\bar{E}_{J_t,*}$, where we have $J_t = \{j: t-100 \le j \le t-1\}$, as the training set. Then for each p and λ , based on the training set $\bar{E}_{J_t,*}$, we calculate the transition matrix estimates $\hat{A}_1^t(p,\lambda), \ldots, \hat{A}_p^t(p,\lambda)$. We then use the obtained estimates to predict the stock price in day t. The averaged prediction error for each specific λ and p is calculated as

$$\overline{\mathrm{Err}}(p,\lambda) = \frac{1}{10} \sum_{t=1}^{10} \|\bar{E}_{t,*} - \sum_{k=1}^{p} \widehat{A}_{k}^{t}(p,\lambda)^{\mathrm{T}} \bar{E}_{t-k,*}\|_{2}.$$

In Table 16, we present the minimized averaged prediction errors $\min_{\lambda} \overline{\operatorname{Err}}(p,\lambda)$ for the three methods with different lag p. The standard deviations of the prediction errors are presented in the parentheses. Our method outperforms the two competing methods in terms of prediction accuracy.

Secondly, we provide the prediction error on day t=1258 based on the selected (p,λ) using cross-validation. By observing Table 16, we select the lag p=1 and the corresponding λ for our method. The prediction error is 7.62 for our method. In comparison, the lasso method and ridge method have the prediction errors 11.11 and 11.94 separately.

lag	ridge method	lasso method	our method
p=1	17.68 (2.49)	15.67(2.74)	11.88 (3.34)
p=2	15.63(3.01)	15.69(2.84)	12.01 (3.41)
p=3	15.17 (3.53)	15.76 (2.83)	12.04 (3.42)
p=4	14.90 (3.69)	15.68 (2.76)	12.02(3.41)
p=5	14.73 (3.66)	$15.62 \ (2.55)$	12.08 (3.29)
p=6	14.58 (3.57)	$15.51 \ (2.58)$	12.09(3.15)
p=7	14.42 (3.49)	15.45 (2.59)	$12.21 \ (3.16)$
p=8	14.36 (3.42)	$15.40 \ (2.57)$	12.25 (3.16)
p=9	14.20 (3.31)	15.28 (2.46)	12.24 (3.06)

Table 16: The optimized averaged prediction errors for the three methods on the equity data, under different lags p from 1 to 9. The standard deviations are present in the parentheses. The smallest prediction error within each column is bolded.

6. Discussions

Estimation of the vector autoregressive model is an interesting problem and has been investigated for a long time. This problem is intrinsically linked to the regression problem with multiple responses. Accordingly (penalized) least squares estimates, which has the maximum likelihood interpretation behind it, look like reasonable solutions. However, high dimensionality brings significantly new challenges and viewpoints to this classic problem. In parallel to the Dantzig selector proposed by Candes and Tao (2007) in cracking the ordinary linear regression model, we advocate borrowing the strength of the linear program in estimating the VAR model. As has been repeatedly stated in the main text, this new formulation brings some advantages over the least square estimates. Moreover, our theoretical analysis brings new insights into the problem of transition matrix estimation, and we highlight the role of $||A_1||_2$ in evaluating the estimation accuracy of the estimator.

In the main text we do not discuss estimating the covariance matrix Σ and Ψ . Lemma 1 builds the $L_{\rm max}$ convergence result for estimating Σ . If we further suppose that the covariance matrix Σ is sparse in some sense, then we can exploit the well developed results in covariance matrix estimation (including "banding", Bickel and Levina, 2008b, "tapering", Cai et al., 2010, and "thresholding", Bickel and Levina, 2008a) to estimate the covariance matrix Σ and establish the consistency result with regard to the matrix L_1 and L_2 norms. With both Σ and A estimated by some constant estimator $\widehat{\Sigma}$, an estimator $\widehat{\Psi}$ of Ψ can be obtained under the VAR model (with lag one) as

$$\widehat{\Psi} = \widehat{\Sigma} - \widehat{A}_1^{\mathrm{T}} \widehat{\Sigma} \widehat{A}_1,$$

and a similar estimator can be built for lag p VAR model using the augmented formulation shown in Equation (4).

In this manuscript we focus on the stationary vector autoregressive model and our method is designed for such stationary process. The stationary requirement is a common assumption in analysis and is adopted by most recent works, for example, Kock and Callot (2015) and Song and Bickel (2011). We notice that there are works in handling unstable VAR models, checking for example Song et al. (2014) and Kock (2012). We would like to explore this problem in the future. Another unexplored region is how to determine the order (lag) of the vector autoregression aside from using the cross-validation approach. There have been results in this area (e.g., Song and Bickel, 2011) and we are also interested in finding whether the linear program can also be exploited in determining the order of the VAR model.

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Appendix A. Proofs of Main Results

In this section we provide the proofs of the main results in the manuscript.

A.1 Proof of Theorem 1

Before proving the main result in Theorem 1, we first establish several lemmas. In the sequel, because we only focus on the lag 1 autoregressive model, for notation simplicity, in $\Sigma_i(\{(X_t)\})$ we remove $\{(X_t)\}$ and simply denote the lag *i* covariance matrix to be Σ_i .

The following lemma describes the L_{max} rate of convergence S to Σ . This result generalizes the upper bound derived when data are independently generated (see, for example, Bickel and Levina, 2008a).

Lemma 1 Letting S be the marginal sample covariance matrix defined in (7), when $T \ge \max(6 \log d, 1)$, we have, with probability no smaller than $1 - 6d^{-1}$,

$$||S - \Sigma||_{\max} \le \frac{16||\Sigma||_2 \max_j(\Sigma_{jj})}{\min_j(\Sigma_{jj})(1 - ||A_1||_2)} \left\{ \left(\frac{6 \log d}{T}\right)^{1/2} + 2\left(\frac{1}{T}\right)^{1/2} \right\}.$$

Proof [**Proof**] For any $j, k \in \{1, 2, ..., d\}$, we have

$$\mathbb{P}(|S_{jk} - \Sigma_{jk}| > \eta) = \mathbb{P}\left(\left|\frac{1}{T}\sum_{t=1}^{T} X_{tj} X_{tk} - \Sigma_{jk}\right| > \eta\right).$$

Letting $Y_t = \{X_{t1}(\Sigma_{11})^{-1/2}, \dots, X_{td}(\Sigma_{dd})^{-1/2}\}^{\mathrm{T}}$ for $t = 1, \dots, T$ and $\rho_{jk} = \Sigma_{jk}(\Sigma_{jj}\Sigma_{kk})^{-1/2}$, we have

$$\mathbb{P}(|S_{jk} - \Sigma_{jk}| > \eta) = \mathbb{P}\left\{ \left| \frac{1}{T} \sum_{t=1}^{T} Y_{tj} Y_{tk} - \rho_{jk} \right| > \eta(\Sigma_{jj} \Sigma_{kk})^{-1/2} \right\} \\
= \mathbb{P}\left\{ \left| \frac{\sum_{t=1}^{T} (Y_{tj} + Y_{tk})^2 - \sum_{t=1}^{T} (Y_{tj} - Y_{tk})^2}{4T} - \rho_{jk} \right| > \eta(\Sigma_{jj} \Sigma_{kk})^{-1/2} \right\} \\
\leq \mathbb{P}\left\{ \left| \frac{1}{T} \sum_{t=1}^{T} (Y_{tj} + Y_{tk})^2 - 2(1 + \rho_{jk}) \right| > 2\eta(\Sigma_{jj} \Sigma_{kk})^{-1/2} \right\} \\
+ \mathbb{P}\left\{ \left| \frac{1}{T} \sum_{t=1}^{T} (Y_{tj} - Y_{tk})^2 - 2(1 - \rho_{jk}) \right| > 2\eta(\Sigma_{jj} \Sigma_{kk})^{-1/2} \right\}. \tag{23}$$

Using the property of Gaussian distribution, we have $(Y_{1j} + Y_{1k}, \dots, Y_{Tj} + Y_{Tk})^T \sim N_T(0, Q)$ for some positive definite matrix Q. In particular, we have

$$\begin{aligned} |Q_{il}| &= |\operatorname{Cov}(Y_{ij} + Y_{ik}, Y_{lj} + Y_{lk})| = |\operatorname{Cov}(Y_{ij}, Y_{lj}) + \operatorname{Cov}(Y_{ij}, Y_{lk}) + \operatorname{Cov}(Y_{ik}, Y_{lk}) + \operatorname{Cov}(Y_{ik}, Y_{lj})| \\ &\leq \frac{1}{\min_{j}(\Sigma_{jj})} |\operatorname{Cov}(X_{ij}, X_{lj}) + \operatorname{Cov}(X_{ij}, X_{lk}) + \operatorname{Cov}(X_{ik}, X_{lk}) + \operatorname{Cov}(X_{ik}, X_{lj})| \\ &\leq \frac{4}{\min_{j}(\Sigma_{jj})} \|\Sigma_{l-i}\|_{\max} \leq \frac{8\|\Sigma\|_{2} \|A_{1}\|_{2}^{|l-i|}}{\min_{j}(\Sigma_{jj})}, \end{aligned}$$

where the last inequality follows from (3).

Therefore, using the matrix norm inequality,

$$||Q||_2 \le \max_{1 \le i \le T} \sum_{l=1}^{T} |Q_{il}| \le \frac{8||\Sigma||_2}{\min_j(\Sigma_{jj})(1 - ||A_1||_2)}.$$

Then applying Lemma 3 to (23), we have

$$\mathbb{P}\left\{ \left| \frac{1}{T} \sum_{t=1}^{T} (Y_{tj} + Y_{tk})^2 - 2(1 + \rho_{jk}) \right| > 2\eta (\Sigma_{jj} \Sigma_{kk})^{-1/2} \right\} \\
\leq 2 \exp\left[-\frac{T}{2} \left\{ \frac{\eta \min_{j} (\Sigma_{jj}) (1 - ||A_1||_2)}{16 ||\Sigma||_2 (\Sigma_{jj} \Sigma_{kk})^{1/2}} - 2T^{-1/2} \right\}^2 \right] + 2 \exp\left(-\frac{T}{2} \right). \tag{24}$$

Using a similar argument, we have

$$\mathbb{P}\left\{ \left| \frac{1}{T} \sum_{t=1}^{T} (Y_{tj} - Y_{tk})^2 - 2(1 - \rho_{jk}) \right| > 2\eta (\Sigma_{jj} \Sigma_{kk})^{-1/2} \right\} \\
\leq 2 \exp \left[-\frac{T}{2} \left\{ \frac{\eta \min_{j} (\Sigma_{jj}) (1 - ||A_1||_2)}{16||\Sigma||_2 (\Sigma_{jj} \Sigma_{kk})^{1/2}} - 2T^{-1/2} \right\}^2 \right] + 2 \exp \left(-\frac{T}{2} \right). \tag{25}$$

Combining (24) and (25), then applying the union bound, we have

$$\mathbb{P}(||S-\Sigma||_{\max}>\eta)$$

$$\leq 3d^2 \exp\left(-\frac{T}{2}\right) + 3d^2 \exp\left[-\frac{T}{2} \left\{ \frac{\eta \min_j(\Sigma_{jj})(1 - ||A_1||_2)}{16||\Sigma||_2 \max_j(\Sigma_{jj})} - 2\left(\frac{1}{T}\right)^{-1/2} \right\}^2\right].$$

The proof thus completes by choosing η as the described form.

In the next lemma we try to quantify the difference between S_1 and Σ_1 with respect to the matrix L_{max} norm. Remind that $\Sigma_1\{(X_t)\}$ is simplified to be Σ_1 .

Lemma 2 Letting S_1 be the lag 1 sample covariance matrix, when $T \ge \max(6 \log d + 1, 2)$, we have, with probability no smaller than $1 - 8d^{-1}$,

$$||S_1 - \Sigma_1||_{\max} \le \frac{32||\Sigma||_2 \max_j(\Sigma_{jj})}{\min_j(\Sigma_{jj})(1 - ||A_1||_2)} \left\{ \left(\frac{3\log d}{T}\right)^{1/2} + \left(\frac{2}{T}\right)^{1/2} \right\}.$$

Proof [**Proof**] We have, for any $j, k \in \{1, 2, ..., d\}$,

$$\mathbb{P}(|(S_1)_{jk} - (\Sigma_1)_{jk}| > \eta) = \mathbb{P}\left(\left|\frac{1}{T-1}\sum_{t=1}^{T-1} X_{tj} X_{(t+1)k} - (\Sigma_1)_{jk}\right| > \eta\right).$$

Letting $Y_t = \{X_{t1}(\Sigma_{11})^{-1/2}, \dots, X_{td}(\Sigma_{dd})^{-1/2}\}^T$ and $\rho_{jk} = (\Sigma_1)_{jk}(\Sigma_{jj}\Sigma_{kk})^{-1/2}$, we have

$$\mathbb{P}(|(S_{1})_{jk} - (\Sigma_{1})_{jk}| > \eta) = \mathbb{P}\left\{ \left| \frac{1}{T-1} \sum_{t=1}^{T-1} Y_{tj} Y_{(t+1)k} - \rho_{jk} \right| > \eta(\Sigma_{jj} \Sigma_{kk})^{-1/2} \right\} \\
= \mathbb{P}\left[\left| \frac{\sum_{t=1}^{T-1} \{Y_{tj} + Y_{(t+1)k}\}^{2} - \sum_{t=1}^{T-1} \{Y_{tj} - Y_{(t+1)k}\}^{2}}{4(T-1)} - \rho_{jk} \right| > \eta(\Sigma_{jj} \Sigma_{kk})^{-1/2} \right] \\
\leq \mathbb{P}\left[\left| \frac{\sum_{t=1}^{T-1} \{Y_{tj} + Y_{(t+1)k}\}^{2}}{T-1} - 2(1 + \rho_{jk}) \right| > 2\eta(\Sigma_{jj} \Sigma_{kk})^{-1/2} \right] \\
+ \mathbb{P}\left[\left| \frac{\sum_{t=1}^{T-1} \{Y_{tj} - Y_{(t+1)k}\}^{2}}{T-1} - 2(1 - \rho_{jk}) \right| > 2\eta(\Sigma_{jj} \Sigma_{kk})^{-1/2} \right]. \tag{26}$$

Using the property of Gaussian distribution, we have $\{Y_{1j} + Y_{2k}, \dots, Y_{(T-1)j} + Y_{Tk}\}^{\mathrm{T}} \sim N_{T-1}(0,Q)$, for some positive definite matrix Q. In particular, we have

$$\begin{split} |Q_{il}| &= |\operatorname{Cov}\{Y_{ij} + Y_{(i+1)k}, Y_{lj} + Y_{(l+1)k}\}| \\ &= |\operatorname{Cov}(Y_{ij}, Y_{lj}) + \operatorname{Cov}\{Y_{ij}, Y_{(l+1)k}\} + \operatorname{Cov}\{Y_{(i+1)k}, Y_{lj}\} + \operatorname{Cov}\{Y_{(i+1)k}, Y_{(l+1)k}\}| \\ &\leq \frac{1}{\min_{j}(\Sigma_{jj})} |\operatorname{Cov}(X_{ij}, X_{lj}) + \operatorname{Cov}\{X_{ij}, X_{(l+1)k}\} + \operatorname{Cov}\{X_{(i+1)k}, X_{lj}\} + \operatorname{Cov}\{X_{(i+1)k}, X_{(l+1)j}\}| \\ &\leq \frac{2\|\Sigma_{l-i}\|_{\max} + \|\Sigma_{l+1-i}\|_{\max} + \|\Sigma_{l-1-i}\|_{\max}}{\min_{j}(\Sigma_{jj})} \\ &\leq \frac{\|\Sigma\|_{2}(2\|A_{1}\|_{2}^{|l-i|} + \|A_{1}\|_{2}^{|l+1-i|} + \|A_{1}\|_{2}^{|l-1-i|})}{\min_{j}(\Sigma_{jj})}. \end{split}$$

Therefore, using the matrix norm inequality,

$$||Q||_2 \le \max_{1 \le i \le (T-1)} \sum_{l=1}^{T-1} |Q_{il}| \le \frac{8||\Sigma||_2}{\min_j(\Sigma_{jj})(1 - ||A_1||_2)}.$$

Then applying Lemma 3 to (26), we have

$$\mathbb{P}\left[\left|\frac{1}{T-1}\sum_{t=1}^{T-1}\left\{Y_{tj}+Y_{(t+1)k}\right\}^{2}-2(1+\rho_{jk})\right|>2\eta(\Sigma_{jj}\Sigma_{kk})^{-1/2}\right]\leq 2\exp\left[-\frac{(T-1)}{2}\left\{\frac{\eta\min_{j}(\Sigma_{jj})(1-\|A_{1}\|_{2})}{16\|\Sigma\|_{2}(\Sigma_{jj}\Sigma_{kk})^{1/2}}-2(T-1)^{-1/2}\right\}^{2}\right]+2\exp\left(-\frac{T-1}{2}\right). \quad (27)$$

Using a similar technique, we have

$$\mathbb{P}\left[\left|\frac{1}{T-1}\sum_{t=1}^{T-1}\left\{Y_{tj}-Y_{(t+1)k}\right\}^{2}-2(1-\rho_{jk})\right|>2\eta(\Sigma_{jj}\Sigma_{kk})^{-1/2}\right]\leq 2\exp\left[-\frac{(T-1)}{2}\left\{\frac{\eta\min_{j}(\Sigma_{jj})(1-\|A_{1}\|_{2})}{16\|\Sigma\|_{2}(\Sigma_{jj}\Sigma_{kk})^{1/2}}-2(T-1)^{-1/2}\right\}^{2}\right]+2\exp\left(-\frac{T-1}{2}\right). \quad (28)$$

Combining (27) and (28), and applying the union bound across all pairs (j, k), we have

$$\mathbb{P}(\|S_1 - \Sigma_1\|_{\max} > \eta) \le 4d^2 \exp\left[-\frac{(T-1)}{2} \left\{ \frac{\eta \min_j(\Sigma_{jj})(1 - \|A_1\|_2)}{16\|\Sigma\|_2 \max_j(\Sigma_{jj})} - 2(T-1)^{-1/2} \right\}^2 \right] + 4d^2 \exp\left(-\frac{T-1}{2}\right).$$

Finally noting that when $T \ge 3$, we have 1/(T-1) < 2/T. The proof thus completes by choosing η as stated.

Using the above two technical lemmas, we can then proceed to the proof of the main results in Theorem 1.

Proof [**Proof of Theorem 1**] With Lemmas 1 and 2, we proceed to prove Theorem 1. We first denote

$$\zeta_{1} = \frac{16\|\Sigma\|_{2} \max_{j}(\Sigma_{jj})}{\min_{j}(\Sigma_{jj})(1 - \|A_{1}\|_{2})} \left\{ \left(\frac{6 \log d}{T}\right)^{1/2} + 2\left(\frac{1}{T}\right)^{1/2} \right\},$$

$$\zeta_{2} = \frac{32\|\Sigma\|_{2} \max_{j}(\Sigma_{jj})}{\min_{j}(\Sigma_{jj})(1 - \|A_{1}\|_{2})} \left\{ \left(\frac{3 \log d}{T}\right)^{1/2} + \left(\frac{2}{T}\right)^{1/2} \right\}.$$

Using Lemmas 1 and 2, we have, with probability no smaller than $1 - 14d^{-1}$,

$$||S - \Sigma||_{\max} \le \zeta_1, \quad ||S_1 - \Sigma_1||_{\max} \le \zeta_2.$$

We firstly prove that population quantity A_1 is a feasible solution to the optimization problem in (10) with probability no smaller than $1 - 14d^{-1}$

$$||SA_{1} - S_{1}||_{\max} = ||S\Sigma^{-1}\Sigma_{1} - S_{1}||_{\max}$$

$$= ||S\Sigma^{-1}\Sigma_{1}^{T} - \Sigma_{1} + \Sigma_{1} - S_{1}||_{\max}$$

$$\leq ||(S\Sigma^{-1} - I_{d})\Sigma_{1}||_{\max} + ||\Sigma_{1} - S_{1}||_{\max}$$

$$\leq ||(S - \Sigma)\Sigma^{-1}\Sigma_{1}||_{\max} + \zeta_{2}$$

$$\leq \zeta_{1}||A_{1}||_{1} + \zeta_{2}$$

$$\leq \lambda_{0}.$$

The last inequality holds by using the condition that $d \geq 8$ implies that $1/T \leq \log d/(2T)$. Therefore, A_1 is feasible in the optimization equation, by checking the equivalence between (10) and (11), we have $\|\widehat{\Omega}\|_1 \leq \|A_1\|_1$ with probability no smaller than $1 - 14d^{-1}$. We then have

$$\begin{split} \|\widehat{\Omega} - A_1\|_{\text{max}} &= \|\widehat{\Omega} - \Sigma^{-1}\Sigma_1\|_{\text{max}} \\ &= \|\Sigma^{-1}(\Sigma\widehat{\Omega} - \Sigma_1)\|_{\text{max}} \\ &= \|\Sigma^{-1}(\Sigma\widehat{\Omega} - S_1 + S_1 - \Sigma_1)\|_{\text{max}} \\ &= \|\Sigma^{-1}(\Sigma\widehat{\Omega} - S\widehat{\Omega} + S\widehat{\Omega} - S_1) + \Sigma^{-1}(S_1 - \Sigma_1)\|_{\text{max}} \\ &\leq \|(I_d - \Sigma^{-1}S)\widehat{\Omega}\|_{\text{max}} + \|\Sigma^{-1}(S\widehat{\Omega} - S_1)\|_{\text{max}} + \|\Sigma^{-1}(S_1 - \Sigma_1)\|_{\text{max}} \\ &\leq \|(I_d - \Sigma^{-1}S)\widehat{\Omega}\|_{\text{max}} + \|\Sigma^{-1}(S_1 - S_1)\|_{\text{max}} + \|\Sigma^{-1}(S_1 - S_1)\|_{\text{max}} \\ &\leq \|\Sigma^{-1}\|_1 \|(\Sigma - S)\widehat{\Omega}\|_{\text{max}} + \|\Sigma^{-1}\|_1 \|S\widehat{\Omega} - S_1\|_{\text{max}} + \|\Sigma^{-1}\|_1 \|S_1 - \Sigma_1\|_{\text{max}} \\ &\leq \|\Sigma^{-1}\|_1 (\|A_1\|_1 \zeta_1 + \lambda_0 + \zeta_2) \\ &= 2\lambda_0 \|\Sigma^{-1}\|_1. \end{split}$$

Let λ_1 be a threshold level and we define

$$s_1 = \max_{1 \le j \le d} \sum_{i=1}^d \min \{ |(A_1)_{ij}|/\lambda_1, 1 \}, \qquad T_j = \{i : |(A_1)_{ij}| \ge \lambda_1 \}.$$

We have, with probability no smaller than $1 - 14d^{-1}$, for all $j \in \{1, \ldots, d\}$,

$$\begin{split} \|\widehat{\Omega}_{*,j} - (A_1)_{*,j}\|_1 &\leq \|\widehat{\Omega}_{T_j^c,j}\|_1 + \|(A_1)_{T_j^c,j}\|_1 + \|\widehat{\Omega}_{T_j,j} - (A_1)_{T_j,j}\|_1 \\ &= \|\widehat{\Omega}_{*,j}\|_1 - \|\widehat{\Omega}_{T_j,j}\|_1 + \|(A_1)_{T_j^c,j}\|_1 + \|\widehat{\Omega}_{T_j,j} - (A_1)_{T_j,j}\|_1 \\ &\leq \|(A_1)_{*,j}\|_1 - \|\widehat{\Omega}_{T_j,j}\|_1 + \|(A_1)_{T_j^c,j}\|_1 + \|\widehat{\Omega}_{T_j,j} - (A_1)_{T_j,j}\|_1 \\ &\leq 2\|(A_1)_{T_j^c,j}\|_1 + 2\|\widehat{\Omega}_{T_j,j} - (A_1)_{T_j,j}\|_1 \\ &\leq 2\|(A_1)_{T_j^c,j}\|_1 + 4\lambda_0\|\Sigma^{-1}\|_1|T_j| \\ &\leq (2\lambda_1 + 4\lambda_0\|\Sigma^{-1}\|_1)s_1. \end{split}$$

Suppose $\max_{j} \sum_{i=1}^{d} |(A_1)_{ij}|^q \le s$ and setting $\lambda_1 = 2\lambda_0 \|\Sigma^{-1}\|_1$, we have

$$\lambda_1 s_1 = \max_{1 \le j \le d} \sum_{i=1}^d \min\{|(A_1)_{ij}|, \lambda_1\} \le \lambda_1 \max_{1 \le j \le d} \sum_{i=1}^d \min\{|(A_1)_{ij}|^q / \lambda_1^q, 1\} \le \lambda_1^{1-q} s.$$

Therefore, we have

$$\|\widehat{\Omega}_{*,j} - (A_1)_{*,j}\|_{1} \le 4\lambda_1 s_1 \le 4\lambda_1^{1-q} s = 4s(2\lambda_0 \|\Sigma^{-1}\|_1)^{1-q}.$$

Noting that when the lag of the time series p=1, by definition in (12), we have $\widehat{\Omega}=\widehat{A}_1$. This completes the proof.

A.2 Proof of the Rest Results

Proof [**Proof of Corollary 1**] Corollary 1 directly follows from Theorem 1, so its proofs is omitted.

Proof [**Proof of Corollary 2**] Using the generating model described in Equation (2), we have

$$||X_{T+1} - \widehat{A}_1^{\mathrm{T}} X_T||_{\infty} = ||(A_1^{\mathrm{T}} - \widehat{A}_1^{\mathrm{T}}) X_T + Z_{T+1}||_{\infty}$$

$$\leq ||A_1^{\mathrm{T}} - \widehat{A}_1^{\mathrm{T}}||_{\infty} ||X_T||_{\infty} + ||Z_{T+1}||_{\infty}$$

$$= ||A_1 - \widehat{A}_1||_1 ||X_T||_{\infty} + ||Z_{T+1}||_{\infty}$$

Using Lemma 4 in Appendix B, we have

$$\mathbb{P}(\|X_T\|_{\infty} \leq (\Sigma_{\max} \cdot \alpha \log d)^{1/2}, \|Z_{T+1}\|_{\infty} \leq (\Psi_{\max} \cdot \alpha \log d)^{1/2}) \geq 1 - 2(d^{\alpha/2 - 1} \sqrt{\pi/2 \cdot \alpha \log d})^{-1}.$$

This, combined with Theorem 1, gives Equation (20).

Proof [Proof of Corollary 3] Similar as the proof in Corollary 2, we have

$$||X_{T+1} - \bar{A}_1^{\mathrm{T}} X_T||_2 = ||(A_1^{\mathrm{T}} - \bar{A}_1^{\mathrm{T}}) X_T + Z_{T+1}||_2$$

$$\leq ||A_1 - \bar{A}_1||_2 ||X_T||_2 + ||Z_{T+1}||_2.$$

For any Gaussian random vector $Y \sim N_d(0,Q)$, we have $Y = \sqrt{Q}Y_0$ where $Y_0 \sim N_d(0,I_d)$. Using the concentration inequality for Lipschitz functions of standard Gaussian random vector (see, for example, Theorem 3.4 in Massart, 2007), we have

$$\mathbb{P}(|||Y||_{2} - E||Y||_{2}| \ge t) = \mathbb{P}(|||\sqrt{Q}Y_{0}||_{2} - E||\sqrt{Q}Y_{0}||_{2}| \ge t)$$

$$\le 2\exp\left(-\frac{t^{2}}{2||Q||_{2}}\right). \tag{29}$$

Here the inequality exploits the fact that for any vectors $x, y \in \mathbb{R}^d$,

$$\|\sqrt{Q}x\|_2 - \|\sqrt{Q}y\|_2 \le \|\sqrt{Q}(x-y)\|_2 \le \|\sqrt{Q}\|_2 \|x-y\|_2$$

and accordingly the function $x \to \|\sqrt{Q}x\|_2$ has the Lipschitz norm no greater than $\sqrt{\|Q\|_2}$. Using Equation (29), we then have

$$\mathbb{P}(\|X_T\|_2 \le \sqrt{2\|\Sigma\|_2 \log d} + E\|X_T\|_2, \ \|Z_{T+1}\|_2 \le \sqrt{2\|\Psi\|_2 \log d} + E\|Z_{T+1}\|_2) \ge 1 - 4d^{-1}.$$

Finally, we have

$$(E||Y||_2)^2 \le E||Y||_2^2 = \text{tr}(Q).$$

Combined with Theorem 1 and the fact that $||A_1 - \bar{A}_1||_2 \le ||A_1 - \bar{A}_1||_1$, we have the desired result.

Appendix B. Supporting Lemmas

Lemma 3 (Negahban and Wainwright, 2011) Suppose that $Y \sim N_T(0, Q)$ is a Gaussian random vector. We have, for $\eta > 2T^{-1/2}$,

$$\mathbb{P}\Big\{ \big| \|Y\|_2^2 - E(\|Y\|_2^2) \big| > 4T\eta \|Q\|_2 \Big\} \le 2 \exp\Big\{ -T(\eta - 2T^{-1/2})^2/2 \Big\} + 2 \exp(-T/2).$$

Proof [**Proof**] This can be proved by first using the concentration inequality for the Lipschitz functions $||Y||_2$ of Gaussian random variables Y. Then combining with the result

$$||Y||_2^2 - E(||Y||_2^2) \le (||Y||_2 - E||Y||_2) \cdot (||Y||_2 + E||Y||_2),$$

we have the desired concentration inequality.

Lemma 4 Suppose that $Z = (Z_1, \ldots, Z_d)^T \in N_d(0, Q)$ is a Gaussian random vector. Letting $Q_{\max} := \max_i(Q_{ii})$, we have

$$\mathbb{P}\{\|Z\|_{\infty} > (Q_{\max} \cdot \alpha \log d)^{1/2}\} \le \left(d^{\alpha/2 - 1} \sqrt{\pi/2 \cdot \alpha \log d}\right)^{-1}.$$

Proof [Proof] Simply using the Gaussian tail probability, we have

$$\mathbb{P}(\|Z\|_{\infty} > t) \leq \sum_{i=1}^{d} \mathbb{P}(|Z_{i}| \cdot Q_{ii}^{-1/2} > t \cdot Q_{ii}^{-1/2}) \leq \sum_{i=1}^{d} \frac{2 \exp(-t^{2}/2Q_{ii})}{t \cdot Q_{ii}^{-1/2} \cdot \sqrt{2\pi}} \leq \frac{2d \exp(-t^{2}/2Q_{\max})}{t \cdot Q_{\max}^{-1/2} \cdot \sqrt{2\pi}}.$$

Taking $t = (Q_{\text{max}} \cdot \alpha \log d)^{1/2}$ into the upper equation, we have the desired result.

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