# NUMERICAL METHODS FOR COMPUTING TWO KINDS OF THE BEST HANKEL TENSOR APPROXIMATIONS* 

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#### Abstract

Hankel tensors and their approximation problems are of particular interest in the multidimensional seismic trace interpolator problem. In this paper, we investigate the numerical methods for two kinds of the best Hankel tensor approximation problems. Based on the Vandermonde decomposition of Hankel tensors, the Hankel tensor approximation problem with missing data is transformed into an unconstrained optimization problem, and then the BFGS method is used to solve it. For the Hankel tensor approximation problems with the interval constraint and box constraint, Dykstra's algorithm and its acceleration versions are designed to solve them. Numerical examples illustrate that these methods are feasible and effective.


Key words: Hankel tensor approximation, BFGS method, Dykstra's algorithm, acceleration technique
Mathematics Subject Classification: 16A69, 65F99

## 1 Introduction

Throughout this paper, to distinguish scalars, vectors, matrices and higher-order tensors, scalars will be denoted by lower case Greek letters, e.g. $\alpha, \beta$, vectors will be denoted by lowercase letters, e.g. $v, w$, matrices will be defined by uppercase letters, e.g. $A, B$ and higher-order tensors will be denoted by calligraphic letters, e.g. $\mathcal{A}, \mathcal{B}$. Denote that $[n]:=$ $\{1,2, \ldots, n\}$. A tensor $\mathcal{A}=\left(a_{i_{1} i_{2} \ldots i_{m}}\right)$ is said to be a symmetric tensor if every entry $a_{i_{1} i_{2} \ldots i_{m}}$ is invariant under any index permutation. Let the vector $v=\left(v_{0}, v_{1}, \ldots, v_{(n-1) m}\right)^{T}$. Define the symmetric tensor $\mathcal{A}=\left(a_{i_{1} i_{2} \ldots i_{m}}\right)$ associated with $v$ by

$$
\begin{equation*}
a_{i_{1} i_{2} \ldots i_{m}}=v_{i_{1}+i_{2}+\cdots+i_{m}-m} \tag{1.1}
\end{equation*}
$$

for $i_{1}, i_{2}, \ldots, i_{m} \in[n]$. Then $\mathcal{A}$ is called a Hankel tensor and $v$ is called the generating vector of $\mathcal{A}$. We see that a sufficient and necessary condition for a symmetric tensor $\mathcal{A}=\left(a_{i_{1} i_{2} \ldots i_{m}}\right)$ to be a Hankel tensor is that whenever $i_{1}+i_{2}+\cdots+i_{m}=j_{1}+j_{2}+\cdots+j_{m}$ we have $a_{i_{1} i_{2} \ldots i_{m}}=a_{j_{1} j_{2} \ldots j_{m}}$. An interesting Theorem from [36, 25] is as follows.

An mth order $n$-dimensional Hankel tensor $\mathcal{T}$ can always be decomposed as

$$
\begin{equation*}
\mathcal{T}=\sum_{j=1}^{r} \lambda_{j} \mu_{j} \circ \mu_{j} \circ \cdots \circ \mu_{j}, \tag{1.2}
\end{equation*}
$$

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where each $\lambda_{j}$ is a scalar and $\mu_{j}=\left(1, a_{j}, a_{j}^{2}, \ldots, a_{j}^{n-1}\right)^{T}$ for some number $a_{j}$. In this case, we have $r \leq(n-1) m+1$.

The decomposition (1.2) is called the Vandermonde decomposition of the Hankel tensor $\mathcal{T}$. We call the minimum value $r$ the Vandermonde $\operatorname{rank}$ of $\mathcal{T}$, and denote it by $\operatorname{Vrank}(\mathcal{T})$ (see [36] for more details). We use $T_{m, n}, S_{m, n}, H_{m, n}$ to denote the set of $m$ th order $n$ dimensional tensors, symmetric tensor and Hankel tensor, respectively. The symbol $\mathcal{A} \odot \mathcal{B}$ means that the entrywise product of the $m t h$ order $n$-dimensional tensors $\mathcal{A}$ and $\mathcal{B}$, that is, $\mathcal{A} \odot \mathcal{B}=\left(a_{i_{1} i_{2} \ldots i_{m}} b_{i_{1} i_{2} \ldots i_{m}}\right)$. The symbol $\operatorname{vec}(\cdot)$ stands for the vectorization operator of a matrix or tensor, and $\operatorname{vec}^{-1}(\cdot)$ stands for its inverse operator. Set $\mathcal{U}=\left(u_{i_{1} i_{2}, \ldots, i_{m}}\right)$ and $\mathcal{V}=$ $\left(v_{i_{1} i_{2}, \ldots, i_{m}}\right), \mathcal{V} \geq \mathcal{U}$ means that $v_{i_{1} i_{2}, \ldots, i_{m}} \geq u_{i_{1} i_{2}, \ldots, i_{m}}$ for all possible $\left(i_{1} i_{2}, \ldots, i_{m}\right)$-entries. We define the Frobenius norm of a tensor $\mathcal{A}=\left(a_{i_{1} i_{2} \ldots i_{m}}\right) \in T_{m, n}$ by

$$
\|\mathcal{A}\|=\left(\sum_{i_{1}, i_{2}, \ldots, i_{m}=1}^{n} a_{i_{1} i_{2} \ldots i_{m}}^{2}\right)^{\frac{1}{2}}
$$

which is induced by the inner product

$$
\langle\mathcal{A}, \mathcal{B}\rangle=\sum_{i_{1}, i_{2}, \ldots, i_{m}=1}^{n} a_{i_{1} i_{2} \ldots i_{m}} b_{i_{1} i_{2} \ldots i_{m}}
$$

where $\mathcal{A}, \mathcal{B} \in T_{m, n}$.
In this paper, we consider the following Hankel tensor approximation problems.
Problem I. Given an mth order n-dimensional tensor $\mathcal{A}$ with missing data, find an mth order $n$-dimensional Hankel tensor $\widehat{\mathcal{X}}$ with $\operatorname{Vrank}(\widehat{\mathcal{X}}) \leq r$ such that

$$
\begin{equation*}
\|W \odot(\mathcal{A}-\widehat{\mathcal{X}})\|^{2}=\min _{\mathcal{X} \in H_{m, n},} \operatorname{Vrank}(\mathcal{X}) \leq r\|W \odot(\mathcal{A}-\mathcal{X})\|^{2} \tag{1.3}
\end{equation*}
$$

where $W=\left(w_{i_{1} i_{2} \ldots i_{m}}\right)$ is a $0-1$ tensor, that is, if the element $a_{i_{1} i_{2} \ldots i_{m}}$ of $\mathcal{A}$ is missing then $w_{i_{1} i_{2} \ldots i_{m}}=0$, otherwise $w_{i_{1} i_{2} \ldots i_{m}}=1$.
Problem II. Given an mth order n-dimensional tensor $\mathcal{A}$, find an mth order $n$-dimensional Hankel tensor $\widehat{\mathcal{X}}$ with structure constraint $\digamma_{1}\left(\right.$ or $\left.\digamma_{2}\right)$ such that

$$
\begin{equation*}
\|\mathcal{A}-\widehat{\mathcal{X}}\|^{2}=\min _{\mathcal{X} \in H_{m, n}, \mathcal{X} \in \digamma_{i}}\|\mathcal{A}-\mathcal{X}\|^{2} \tag{1.4}
\end{equation*}
$$

Here the structure constraints are

$$
\begin{gathered}
\digamma_{1}=\left\{\mathcal{X} \in T_{m, n} \mid \quad \mathcal{U} \leq \mathcal{X} \leq \mathcal{V}, \quad \mathcal{U}, \mathcal{V} \in T_{m, n}\right\} \\
\digamma_{2}=\left\{\mathcal{X} \in T_{m, n} \mid \quad\|\mathcal{X}\| \leq \delta, \quad \delta>0\right\}
\end{gathered}
$$

Problem I arises in the multidimensional seismic trace interpolator problem (see [33,21] for more details), which can be stated as follows. Given a raw frequency slice $S$ having two spatial dimensions with lengths $s_{1}$ and $s_{2}$, we can form a fourth-order Hankel tensor $\mathcal{A}$ by generating two tensor orders for every spatial dimension

$$
\mathcal{A}(i, j, m, n)=S(i+j-1, m+n-1)
$$

In four spatial dimensions we build an eighth-order tensor

$$
\mathcal{A}(i, j, m, n, p, q, r, s)=S(i+j-1, m+n-1, p+q-1, r+s-1)
$$

Of course, some elements of the tensor $\mathcal{A}$ will be unknown (and thus in need of interpolating) due to the missing traces. After that we must find a low rank Hankel tensor $\mathcal{X}$, which fits as closely as possible the known elements of $\mathcal{A}$. This process can be summarized as Problem I.

Problem II is of particular interests in the signal processing, such as in the multidimensional harmonic retrieval problems as follows (see [7,11,14,20,24,30,35]). Let us consider $N$ samples of a time series $x_{n}, n \in\{1,2, \ldots, N\}$ modeled as a finite sum of $K$ exponentially damped complex sinusoids

$$
x_{n}=\sum_{p=1}^{P} a_{p} e^{j \phi(p)} e^{\left(-d_{p}+j w_{p}\right) t_{n}}
$$

where $a_{p}$ are the amplitudes, $\phi_{p}$ the phases, $d_{p}$ the damping factors and $w_{p}$ the pulsations. $t_{n}=n \Delta t$ is the time lapse between the time origin and the sample $x_{n}$ and $\Delta t$ is the sampling time interval. These data may be arranged in a high-order Hankel tensor. In order to estimate a tensor data from noisy observations and increase the relative signal strength, we must use a Hankel tensor to approximate a data tensor, in particular, some structured conditions, such as interval constraint and box constraint, are added in the best Hankel tensor approximation problems, which lead to Problem II.

Recently, Hankel tensors have been investigated extensively and the research results mainly concentrated on the spectral characterizations ([25]), tensor decompositions ( $[7,8,10]$ ) and their positivities (including positive definiteness, copositivity, and complete positivity [26-29,36]) and their applications ([11,13]).

When the order $m=2$, the Hankel tensor approximation problem reduces to the Hankel matrix approximation problem. Macinnes [23] developed a method for finding the best approximation of a matrix by a full rank Hankel matrix. The Hankel matrix approximation problem is transformed into a problem involving best approximation of a given vector by a second vector whose elements are constrained so that its inverse image can be algebraically illustrated by Hankel matrix. Based on the projection algorithm and Newton method, AlHomidan [3] proposed a hybrid method for approximating the positive semidefinite Hankel matrix. Arnab, Pokala and Kumaresan [5] presented the algorithms for the least-squares approximation of Toeplitz and Hankel matrices from noise corrupted or ill-composed matrices. The similar structured matrix approximation problems, such as Toeplitz matrix approximation $[1,5,17]$ and positive semidefinite matrix approximation [ 15,31 ], have also been studied, and some excellent numerical methods were developed to solve them. However, the research results of the high-order Hankel tensor approximation problems I and II are very few as far as we know. The biggest obstacle the characterization of the feasible sets of Problems I and II because they have highly structured constraints.

In this paper, we overcome these difficulties by the Vandermonde decomposition and projections. For the problem I, we first transform Problem I into a smooth unconstrained optimization problem by using the Vandermonde decomposition of Hankel tensor, then use the BFGS method with Armijo line search to solve it. For the problem II, we first compute the projections of a tensor onto the sets $H_{m, n}, \digamma_{1}$ and $\digamma_{2}$, and then design the Dykstra's algorithm and its acceleration versions to solve it. Numerical experiments illustrate that the new methods are feasible and effective.

This paper is organized as follows. In Section 2, we first use the Vandermonde decomposition of Hankel tensor to characterize the feasible set of Problem I and then use the BFGS method to solve it. In Section 3, we first compute the projections of a tensor onto the sets $H_{m, n}, \digamma_{1}$ and $\digamma_{2}$, and then design the Dykstra's algorithm and its acceleration versions to
solve Problem II. Finally, some numerical examples are reported to verify the feasibility and effectiveness of these methods.

## 2 BFGS Method for Solving Problem I

In this section, we first transform Problem I into an unconstrained optimization problem, and then use the BFGS method with the Armijo line search to solve it.

Set

$$
\Omega=\left\{\mathcal{X} \in T_{m, n} \mid \quad \mathcal{X} \in H_{m, n}, \operatorname{Vrank}(\mathcal{X}) \leq r\right\}
$$

then from (1.2) it follows that the feasible set $\Omega$ of Problem I can be characterized by

$$
\mathcal{X}=\sum_{j=1}^{r} \lambda_{j} \mu_{j} \circ \mu_{j} \circ \ldots \mu_{j}
$$

where $\mu_{j}=\left(1, x_{j}, \ldots, x_{j}^{n-1}\right)^{T}$. It is easy to verify that

$$
\mathcal{X}=\left(x_{i_{1} i_{2} \ldots i_{m}}\right), \quad x_{i_{1} i_{2} \ldots i_{m}}=\sum_{j=1}^{r} \lambda_{j} x_{j}^{i_{1}+i_{2}+\cdots+i_{m}-m}
$$

Therefore Problem I can also be reformulated as an unconstrained optimization problem

$$
\begin{equation*}
\min _{\eta=\left(x_{1}, x_{2}, \ldots, x_{r}, \lambda_{1}, \lambda_{2}, \ldots, \lambda_{r}\right) \in R^{2 r}} f(\eta), \tag{2.1}
\end{equation*}
$$

where the objective function

$$
\begin{aligned}
f(\eta) & =\|W \odot(\mathcal{A}-\mathcal{X})\|^{2} \\
& =\left\|W \odot\left(\mathcal{A}-\sum_{j=1}^{r} \lambda_{j} \mu_{j} \circ \mu_{j} \circ \ldots \mu_{j}\right)\right\|^{2} \\
& =\sum_{i_{1}, i_{2}, \ldots, i_{m}=1}^{n}\left[w_{i_{1} i_{2} \ldots i_{m}}\left(a_{i_{1} i_{2} \ldots i_{m}}-\sum_{j=1}^{r} \lambda_{j} x_{j}^{i_{1}+i_{2}+\cdots+i_{m}-m}\right)\right]^{2}
\end{aligned}
$$

Now we begin to use the BFGS method with Armijo line search to solve (2.1). The important step to use this method is how to compute the gradient of the objective function $f(\eta)$. Set $p=i_{1}+i_{2}+\cdots+i_{m}-m$, then

$$
\begin{gather*}
\frac{\partial f}{\partial x_{l}}=2 \sum_{i_{1}, i_{2}, \ldots, i_{m}=1}^{n} w_{i_{1} i_{2} \ldots i_{m}}\left(\sum_{j=1}^{s} \lambda_{j} x_{j}^{p}-a_{i_{1} i_{2} \ldots i_{m}}\right)\left(p \lambda_{l} x_{l}^{p-1}\right), \quad l=1,2, \ldots, r  \tag{2.2}\\
\frac{\partial f}{\partial \lambda_{l}}=2 \sum_{i_{1}, i_{2}, \ldots, i_{m}=1}^{n} w_{i_{1} i_{2} \ldots i_{m}}\left(\sum_{j=1}^{s} \lambda_{j} x_{j}^{p}-a_{i_{1} i_{2} \ldots i_{m}}\right)\left(x_{l}^{p}\right), \quad l=1,2, \ldots, r . \tag{2.3}
\end{gather*}
$$

Therefore the gradient

$$
\nabla f=\left(\frac{\partial f}{\partial x_{1}}, \frac{\partial f}{\partial x_{2}}, \ldots, \frac{\partial f}{\partial x_{r}}, \frac{\partial f}{\partial \lambda_{1}}, \frac{\partial f}{\partial \lambda_{2}}, \ldots, \frac{\partial f}{\partial \lambda_{r}}\right)
$$

is given by (2.2) and (2.3).
The BFGS method with Armijo line search to solve (2.1) can be stated as follows.

Algorithm 2.1 (This algorithm attempts to solve (2.1)).

1. Given an initial value $\eta_{0} \in R^{2 r}$, the initial positive definite matrix $B_{0}$ and the tolerant error $\varepsilon>0$. Set $\rho \in(0,1), \sigma \in(0,0.5), k=0$.
2. Compute $\nabla g_{k}=\nabla f\left(\eta_{k}\right)$. If $\left\|\nabla g_{k}\right\| \leq \varepsilon$, then stop and declare that $\eta_{k}$ is a stationary point.
3. Determine $d_{k}$ by solving the systems of linear equations $B_{k} d=-\nabla g_{k}$.
4. Determine the step length $\alpha_{k}$ by Armijo line search, that is, find the smallest nonnegative integer $m_{k}$ such that

$$
f\left(\eta_{k+1}\right) \leq f\left(\eta_{k}\right)+\sigma \rho^{m_{k}} \nabla g_{k}^{T} d_{k}
$$

5. Set $\alpha_{k}=\rho^{m_{k}}, \eta_{k+1}=\eta_{k}+\alpha_{k} d_{k}$.
6. Compute $s_{k}=\eta_{k+1}-\eta_{k}, y_{k}=\nabla g_{k+1}-\nabla g_{k}$.
7. Determine $B_{k+1}$ by

$$
B_{k+1}=\left\{\begin{array}{l}
B_{k}, \quad \text { if } y_{k}^{T} s_{k} \leq 0 \\
B_{k}-\frac{B_{k} s_{k} s_{k}^{T} B_{k}}{s_{k}^{T} B_{k} s_{k}}+\frac{y_{k} y_{k}^{T}}{y_{k}^{T} s_{k}}, \quad \text { if } y_{k}^{T} s_{k}>0
\end{array}\right.
$$

8. Let $k=k+1$. Go to step 2 .

The global convergence theorem for Algorithm 2.1 can be seen in Theorem 10.3.7 of Chen ([11]).

## 3 Dykstra's Algorithm and its Acceleration Versions for Solving Problem II

In this section, we use Dykstra's algorithm and its acceleration versions to compute the Hankel tensor approximation with structure constraints $\digamma_{1}$ and $\digamma_{2}$. Since the objective function of Problem II is a convex function and the feasible set of Problem II is a convex set, so the problem II always has a solution. We first introduce Dykstra's algorithm and its convergence theorem. We begin with a definition.
Definition 3.1 ([6]). Let $M$ be a closed convex subset in a real Hilbert space $H$ and $u$ be a point in $H$, then the unique solution of the following minimization problem

$$
\min _{x \in M}\|x-u\|
$$

is called the projection of $u$ onto $M$ and denoted by $P_{M}(u)$.
In order to find the projection of a given point onto the intersection of a finite number of closed convex sets $C_{1}, C_{2}, \ldots, C_{n}$, Dykstra's algorithm was proposed in [6] which can be stated as follows. This algorithm can be also seen in [12,22].

Dykstra's Algorithm

1. Given the initial value $x_{0}$;
2. Set $x_{n}^{(0)}=x_{0}, \quad I_{i}^{(0)}=0, \quad i=1,2, \ldots, n$.
3. For $k=1,2,3, \ldots$

$$
\begin{aligned}
& x_{0}^{(k)}=x_{n}^{(k-1)} \\
& \quad \text { For } i=1,2, \ldots, n \\
& x_{i}^{(k)}=P_{C_{i}}\left(x_{i-1}^{(k)}-I_{i}^{(k-1)}\right), \\
& I_{i}^{(k)}=x_{i}^{(k)}-\left(x_{i-1}^{(k)}-I_{i}^{(k-1)}\right) . \\
& \quad \text { End }
\end{aligned}
$$

End
The utility of Dykstra's algorithm is based on the following lemma.

Lemma 3.2 ([9], Theorem 2). Let $C_{1}, C_{2}, \ldots, C_{n}$ be some closed convex subsets of a real Hilbert space $H$ such that $C_{1} \bigcap C_{2} \bigcap \cdots \bigcap C_{n} \neq \varnothing$. For any $i=1,2, \ldots, n$ and any $x_{0} \in H$, then the sequences $\left\{x_{i}^{(k)}\right\}$ generated by Dykstra's algorithm converge to $P_{C_{1}} \cap C_{1} \cap \cdots \cap C_{n}\left(x_{0}\right)$, that is,

$$
x_{i}^{(k)} \rightarrow P_{C_{1} \cap C_{1} \cap \cdots \cap C_{n}}\left(x_{0}\right), \quad i=1,2, \ldots, n, \quad k \rightarrow+\infty .
$$

However, the rate of convergence of Dykstra's algorithm is at best linear and it can require a large number of iterations to converge within a given tolerance. In order to accelerate the convergence, the acceleration versions of Dykstra's algorithm were proposed (see [16,19] for more details). We first introduce the Anderson acceleration of the fixed point iteration. Anderson acceleration was proposed by Anderson [4] and developed by Toth and Kelley [32] and Walker and Ni [34]. Recently Higham [16] used this method to compute the nearest correlation matrix problem. Suppose we want to solve the equation $x=f(x)$ for some $g: R^{n} \rightarrow R^{n}$. Basic fixed point iteration for this problem is as follows

$$
\text { Given } x_{0}, \quad x_{k+1}=f\left(x_{k}\right), \quad k=0,1,2, \ldots
$$

For this iteration, the Anderson acceleration is as follows. This algorithm starts with $x_{0} \in R^{n}$ and an integer $m \geq 1$. The following notation is used: $m k=\min (m, k), \triangle x_{i}=x_{i+1}-$ $x_{i}, X_{k}=\left[\triangle x_{k-m k} \ldots \triangle x_{k-1}\right], f_{i}=f\left(x_{i}\right), \triangle f_{i}=f_{i+1}-f_{i}$ and $F_{k}=\left[\triangle f_{k-m k} \ldots \triangle f_{k-1}\right]$.

Anderson acceleration algorithm

1. $x_{1}=x_{0}+f\left(x_{0}\right)$
2. For $k=1,2,3, \ldots$
3. $\quad m k=\min (m, k)$;
4. Determine $\gamma_{k}=\left(\gamma_{k-m k}^{(k)}, \ldots, \gamma_{k-1}^{(k)}\right)^{T}$ that solves $\min _{\gamma \in R^{m k}}\left\|f_{k}-F_{k} \gamma\right\|_{2}^{2}$ by using $Q R$ factorization;
5. $\quad \bar{x}_{k}=x_{k}-\sum_{i=k-m k}^{k-1} \gamma_{i}^{k} \triangle x_{i}=x_{k}-X_{k} \gamma^{(k)}$;
6. $\bar{f}_{k}=f_{k}-\sum_{i=k-m k}^{k-1} \gamma_{i}^{k} \triangle f_{i}=f_{k}-F_{k} \gamma^{(k)}$;
7. $x_{k+1}=\bar{x}+\bar{f}$.
8. End

Now we begin to solve Problem II with structure constraints $\digamma_{1}$ and $\digamma_{2}$. It is easy to verify that Problem II is equivalent to finding the projections $P_{\digamma_{i}} \cap_{H_{m, n}}(\mathcal{A})$ of $\mathcal{A}$ onto the intersection set $\digamma_{i} \bigcap H_{m, n}$. Now we will use Dykstra's algorithm and its acceleration versions to solve Problem II. The key problems to realize these algorithms are how to compute the projections $P_{H_{m, n}}(\mathcal{Z})$ and $P_{\digamma_{i}}(\mathcal{Z}), i=1,2$ of a given tensor $\mathcal{Z}$ onto the sets $H_{m, n}$ and $\digamma_{i}$, respectively. Such problems are perfectly solved as follows.

We can see that $H_{m, n}$ is a subspace of $S_{m, n}$, and its dimension is $(n-1) m+1$. Let $\mathcal{G}_{0}, \mathcal{G}_{1}, \ldots, \mathcal{G}_{(n-1) m}$ be the orthogonal basis of $H_{m, n}$, that is, $\left\langle\mathcal{G}_{i}, \mathcal{G}_{j}\right\rangle=0,0 \leq i \neq j \leq$ $(n-1) m$. Then we have the following results.

Theorem 3.3. For a given $m$ th order n-dimensional tensor $\mathcal{Z}$, then the projection $P_{H_{m, n}}(\mathcal{Z})$ of $\mathcal{Z}$ onto $H_{m, n}$ is

$$
P_{H_{m, n}}(\mathcal{Z})=\sum_{k=0}^{(n-1) m} x_{k} \mathcal{G}_{k}
$$

where

$$
x_{k}=\frac{\left\langle\mathcal{Z}, \mathcal{G}_{k}\right\rangle}{\left\langle\mathcal{G}_{k}, \mathcal{G}_{k}\right\rangle}, \quad k=0,1, \ldots,(n-1) m .
$$

Proof. Since $\mathcal{G}_{0}, \mathcal{G}_{1}, \ldots, \mathcal{G}_{(n-1) m}$ be the orthogonal basis of $H_{m, n}$, then for arbitrary $\mathcal{X} \in$ $H_{m, n} \subset S_{m, n}$, there exists real numbers $x_{0}, x_{1}, \ldots, x_{(n-1) m}$, such that $\mathcal{X}=\sum_{k=0}^{(n-1) m} x_{k} \mathcal{G}_{k}$. Therefore,

$$
\begin{aligned}
f(\mathcal{X}) & =\|\mathcal{Z}-\mathcal{X}\|^{2} \\
& =\left\|\mathcal{Z}-\sum_{(n=0}^{(n-1) m} x_{k} \mathcal{G}_{k}\right\|^{2} \\
& =\left\langle\mathcal{Z}-\sum_{k=0}^{(n-1) m} x_{k} \mathcal{G}_{k}, \mathcal{Z}-\sum_{k=0}^{(n-1) m} x_{k} \mathcal{G}_{k}\right\rangle \\
& =\langle\mathcal{Z}, \mathcal{Z}\rangle-2\left\langle\mathcal{Z}, \sum_{k=0}^{(n-1) m} x_{k} \mathcal{G}_{k}\right\rangle+\left\langle\sum_{k=0}^{(n-1) m} x_{k} \mathcal{G}_{k}, \sum_{k=0}^{(n-1) m} x_{k} \mathcal{G}_{k}\right\rangle \\
& =\|\mathcal{Z}\|^{2}-2 \sum_{k=0}^{(n-1) m} x_{k}\left\langle\mathcal{Z}, \mathcal{G}_{k}\right\rangle+\sum_{k=0}^{(n-1) m} x_{k}^{2}\left\langle\mathcal{G}_{k}, \mathcal{G}_{k}\right\rangle .
\end{aligned}
$$

Set

$$
\frac{\partial f}{\partial x_{i}}=2 x_{k}\left\langle\mathcal{G}_{k}, \mathcal{G}_{k}\right\rangle-2\left\langle\mathcal{Z}, \mathcal{G}_{k}\right\rangle=0
$$

then we derive the stationary points of $f$ are as follows

$$
x_{k}=\frac{\left\langle\mathcal{Z}, \mathcal{G}_{k}\right\rangle}{\left\langle\mathcal{G}_{k}, \mathcal{G}_{k}\right\rangle}, \quad k=0,1, \ldots,(n-1) m
$$

Noting that

$$
\frac{\partial^{2} f}{\partial x_{i}^{2}}=2\left\langle\mathcal{G}_{k}, \mathcal{G}_{k}\right\rangle=2\left\|\mathcal{G}_{k}\right\|^{2}>0
$$

then by Definition 3.1 we obtain that the projection $P_{H_{m, n}}(\mathcal{Z})$ is

$$
P_{H_{m, n}}(\mathcal{Z})=\sum_{k=0}^{(n-1) m} x_{k} \mathcal{G}_{k}
$$

where

$$
x_{k}=\frac{\left\langle\mathcal{Z}, \mathcal{G}_{k}\right\rangle}{\left\langle\mathcal{G}_{k}, \mathcal{G}_{k}\right\rangle}, \quad k=0,1, \ldots,(n-1) m
$$

Proof. Set the $m$ th order $n$-dimensional tensors $\mathcal{V}=\left(v_{i_{1} i_{2}, \ldots, i_{m}}\right)$ and $\mathcal{U}=\left(u_{i_{1} i_{2}, \ldots, i_{m}}\right)$. Suppose that the set $\digamma_{1}$ be nonempty. For a given $m$ th order $n$-dimensional tensor $\mathcal{Z}=$ $\left(z_{i_{1} i_{2}, \ldots, i_{m}}\right)$, then the $\left(i_{1} i_{2}, \ldots, i_{m}\right)$-entry of the projection $P_{\digamma_{1}}(\mathcal{Z})$ is as follows

$$
\left(P_{\digamma_{1}}(\mathcal{Z})\right)_{i_{1} i_{2}, \ldots, i_{m}}=\left\{\begin{array}{lll}
z_{i_{1} i_{2}, \ldots, i_{m}}, & \text { if } u_{i_{1} i_{2}, \ldots, i_{m}} \leq z_{i_{1} i_{2}, \ldots, i_{m}} \leq v_{i_{1} i_{2}, \ldots, i_{m}} \\
v_{i_{1} i_{2}, \ldots, i_{m}}, & \text { if } z_{i_{1} i_{2}, \ldots, i_{m}}>v_{i_{1} i_{2}, \ldots, i_{m}} \\
u_{i_{1} i_{2}, \ldots, i_{m}}, & \text { if } & z_{i_{1} i_{2}, \ldots, i_{m}}<u_{i_{1} i_{2}, \ldots, i_{m}}
\end{array}\right.
$$

Proof. By Definition 3.1 we obtain that the projection $P_{\digamma_{1}}(\mathcal{Z})$ is the unique solution of the following minimization problem

$$
\min _{\mathcal{X} \in F_{1}}\|\mathcal{Z}-\mathcal{X}\|^{2}=\sum_{i_{1}, i_{2}, \ldots, i_{m}=1}^{n}\left(z_{i_{1} i_{2}, \ldots, i_{m}}-x_{i_{1} i_{2}, \ldots, i_{m}}\right)^{2}
$$

This problem attains its minimum if and only if

$$
x_{i_{1} i_{2}, \ldots, i_{m}}= \begin{cases}z_{i_{1} i_{2}, \ldots, i_{m}}, & \text { if } u_{i_{1} i_{2}, \ldots, i_{m}} \leq z_{i_{1} i_{2}, \ldots, i_{m}} \leq v_{i_{1} i_{2}, \ldots, i_{m}} \\ v_{i_{1} i_{2}, \ldots, i_{m}}, & \text { if } z_{i_{1} i_{2}, \ldots, i_{m}}>v_{i_{1} i_{2}, \ldots, i_{m}} \\ u_{i_{1} i_{2}, \ldots, i_{m}}, & \text { if } z_{i_{1} i_{2}, \ldots, i_{m}}<u_{i_{1} i_{2}, \ldots, i_{m}}\end{cases}
$$

which is the projection $P_{\digamma_{1}}(\mathcal{Z})$ of $\mathcal{Z}$ onto the set $\digamma_{1}$.
Theorem 3.4. For a given mth order n-dimensional tensor $\mathcal{Z}=\left(z_{i_{1} i_{2}, \ldots, i_{m}}\right)$, then the projection $P_{\digamma_{2}}(\mathcal{Z})$ of $\mathcal{Z}$ onto $\digamma_{2}$ is as follows

Proof. Set $f(\mathcal{X})=\|\mathcal{Z}-\mathcal{X}\|$. By Definition 3.1 we know that the projection $P_{\digamma_{2}}(Z)$ is the unique solution of the minimization problem $\min _{\mathcal{X} \in \digamma_{2}} f(\mathcal{X})$. It is easy to see that if $\|\mathcal{Z}\|<\delta$, then the unique solution of $\min _{\mathcal{X} \in \digamma_{2}} f(\mathcal{X})$ is $Z$. Now we consider the case $\|\mathcal{Z}\| \geq \delta$. For arbitrary $\mathcal{X} \in \digamma_{2}$ with $\|\mathcal{Z}\| \geq \delta$, then

$$
\begin{aligned}
\|\mathcal{Z}\| & =\|\mathcal{Z}-\mathcal{X}+\mathcal{X}\| \\
& \leq\|\mathcal{Z}-\mathcal{X}\|+\|\mathcal{X}\|
\end{aligned}
$$

which implies that

$$
\begin{aligned}
\|\mathcal{Z}-\mathcal{X}\| & \geq\|\mathcal{Z}\|-\|\mathcal{X}\| \\
& \geq\|\mathcal{Z}\|-\delta
\end{aligned}
$$

This inequality shows that for arbitrary $\mathcal{X} \in \digamma_{2}$ with $\|\mathcal{Z}\| \geq \delta$, we have

$$
\begin{aligned}
f(\mathcal{X}) & =\|\mathcal{Z}-\mathcal{X}\| \\
& \geq\|\mathcal{Z}\|-\delta \\
& =\frac{1}{\|\mathcal{Z}\|}(\|\mathcal{Z}\|-\delta)\|\mathcal{Z}\| \\
& =\frac{1}{\|\mathcal{Z}\|}\| \| \mathcal{Z}\|\mathcal{Z}-\delta \mathcal{Z}\| \\
& =\left\|\mathcal{Z}-\frac{\delta}{\|\mathcal{Z}\|} \mathcal{Z}\right\| \\
& =f\left(\frac{\delta}{\|\mathcal{Z}\|}\right)
\end{aligned}
$$

which shows that when $\mathcal{X}=\frac{\delta}{\|\mathcal{Z}\|} \mathcal{Z}, f(\mathcal{X})$ attains its minimum. Therefore the projection $P_{\digamma_{2}}(\mathcal{Z})$ of $\mathcal{Z}$ onto $\digamma_{2}$ is

By Dykstra's algorithm and combining Theorems 3.3-3.4, we derive a new algorithm to solve Problem II as follows.

Algorithm 3.5 (Dykstra's algorithm for solving Problem II).

1. Set the initial value $\mathcal{X}_{0}=\mathcal{A}$;
2. Set $\mathcal{X}_{2}^{(0)}=X_{0}, \quad \mathcal{I}_{1}^{(0)}=0, \quad \mathcal{I}_{2}^{(0)}=0$;
3. For $k=1,2,3, \ldots$

$$
\mathcal{X}_{0}^{(k)}=\mathcal{X}_{2}^{(k-1)}
$$

$$
\begin{aligned}
& \mathcal{X}_{1}^{(k)}=P_{H_{m, n}}\left(\mathcal{X}_{0}^{(k)}-\mathcal{I}_{1}^{(k-1)}\right) \\
& \mathcal{I}_{1}^{(k)}=\mathcal{X}_{1}^{(k)}-\left(\mathcal{X}_{0}^{(k)}-\mathcal{I}_{1}^{(k-1)}\right), \\
& \mathcal{X}_{2}^{(k)}=P_{F_{i}}\left(\mathcal{X}_{1}^{(k)}-\mathcal{I}_{2}^{(k-1)}\right) \\
& \mathcal{I}_{2}^{(k)}=\mathcal{X}_{2}^{(k)}-\left(\mathcal{X}_{1}^{(k)}-\mathcal{I}_{2}^{(k-1)}\right),
\end{aligned}
$$

End
By Lemma 3.2, we get the convergence theorem for Algorithm 3.1.
Theorem 3.6. If the set $\digamma_{i} \bigcap H_{m, n}$ is nonempty, then the sequences $\left\{\mathcal{X}_{1}^{(k)}\right\}$ and $\left\{\mathcal{X}_{2}^{(k)}\right\}$ generated by Algorithm 3.1 converge to the projection $P_{\digamma_{i} \cap H_{m, n}}(\mathcal{A})$, which is the unique solution of Problem II.

In order to accelerate the convergence of Dykstra's algorithm, the Anderson acceleration algorithm was used in Algorithm 3.1, that is, we will design Anderson acceleration of Dykstra's algorithm to solve Problem II. Set $\Delta \mathcal{S}^{(0)}=0, \mathcal{X}_{2}^{(0)}=\mathcal{A}$ and define the function

$$
\begin{equation*}
\left[\mathcal{X}_{1}^{(k)}, \mathcal{X}_{2}^{(k)}, \Delta \mathcal{S}^{(k)}\right]=g\left(\mathcal{X}_{2}^{(k-1)}, \Delta \mathcal{S}^{(k-1)}\right) \tag{3.2}
\end{equation*}
$$

where

$$
\begin{aligned}
& \mathcal{R}^{(k)}=\mathcal{X}_{2}^{(k-1)}-\Delta \mathcal{S}^{(k-1)} \\
& \mathcal{X}_{1}^{(k)}=P_{H_{m, n}}\left(\mathcal{R}^{(k)}\right) \\
& \Delta \mathcal{S}^{(k)}=\mathcal{X}_{1}^{(k)}-\mathcal{R}^{(k)} \\
& \mathcal{X}_{2}^{(k)}=P_{F_{i}}\left(\mathcal{X}_{1}^{(k)}\right)
\end{aligned}
$$

Then Algorithm 3.1 can be recast as a fixed point method (see [16] for more details)

$$
\left[\mathcal{X}_{1}^{(k)}, \mathcal{X}_{2}^{(k)}, \triangle \mathcal{S}^{(k)}\right]=g\left(\mathcal{X}_{2}^{(k-1)}, \Delta \mathcal{S}^{(k-1)}\right), \quad k=0,1,2, \ldots
$$

Combining Anderson acceleration algorithm, the Anderson acceleration of Dykstra's algorithm to solve Problem II is as follows.

Algorithm 3.7 (Anderson acceleration of Dykstra's algorithm for solving Problem II).

1. Run Anderson acceleration algorithm on the function $f: R^{2 n^{m}} \rightarrow R^{2 n^{m}}$ given by $f(z)=$ $\operatorname{vec}(\tilde{g}(\mathcal{Z})-\mathcal{Z})$, where $z_{k}=\operatorname{vec}\left(\mathcal{Z}_{k}\right), \mathcal{Z}_{k}=\left(\mathcal{X}_{2}^{(k)}, \Delta \mathcal{S}^{(k)}\right) \in R^{n \times 2 n \times \cdots \times n}$ and $\left[\mathcal{X}_{1}^{(k)}, \tilde{g}\left(\mathcal{Z}_{k}\right)\right]=$ $g\left(\mathcal{Z}_{k}\right)$ for the function $g$ defined by (3.2). Terminate the iteration when the stopping criteria attains. Denote the result by $\mathcal{X}_{*}$.
2. Output the solution $\widehat{\mathcal{X}}=\operatorname{vec}^{-1}\left(\mathcal{X}_{*}\right)$.

It is worthy to point out that, unlike Algorithm 3.1, Algorithm 3.7 is not guaranteed to converge, since there are no suitable convergence results for Anderson acceleration. Whether convergence can be proved under reasonable assumptions is an open question.

We begin to discuss the other acceleration scheme for Dykstra's Algorithm (see [19] for more details). For that we need to consider an auxiliary sequence in the product space $H \times H$ that will be denoted as $H^{2}$. For each $k \geq 0$, let's define

$$
\hat{\mathcal{X}}_{k}=\left(\mathcal{X}_{1}^{(k)}, \mathcal{X}_{2}^{(k)}\right) \in H^{2}
$$

where $\mathcal{X}_{1}^{(k)}$ and $\mathcal{X}_{2}^{(k)}$ are defined in Dykstra's Algorithm. The idea to accelerate the convergence of Dykstra's Algorithm is building another sequence in $H^{2}$ which also converges to $P_{C_{1} \cap C_{1} \cap \cdots \cap C_{n}}\left(\mathcal{X}_{0}\right)$. Based on this idea, Lopez and Raydan [19] designed an acceleration scheme of Dykstra's algorithm is as follows

$$
\begin{align*}
& Z_{k}=  \tag{3.3}\\
& \left\{\begin{array}{c}
\frac{1}{2}\left(\mathcal{X}_{2}^{(k+1)}+\mathcal{X}_{1}^{(k+1)}\right), \text { if }\left|\left\|\mathcal{X}_{2}^{(k)}-\mathcal{X}_{1}^{(k)}\right\|+<\mathcal{X}_{2}^{(k)}-\mathcal{X}_{1}^{(k)}, \mathcal{X}_{1}^{(k+1)}-X_{2}^{(k+1)}>\right|<\varepsilon \\
\mathcal{X}_{2}^{(k)}+\alpha_{k}\left(\mathcal{X}_{2}^{(k+1)}-\mathcal{X}_{2}^{(k)}\right), \text { if }\left|\left\|\mathcal{X}_{2}^{(k)}-\mathcal{X}_{1}^{(k)}\right\|+<\mathcal{X}_{2}^{(k)}-\mathcal{X}_{1}^{(k)}, \mathcal{X}_{1}^{(k+1)}-X_{2}^{(k+1)}>\right| \geq \varepsilon
\end{array}\right.
\end{align*}
$$

where $\varepsilon>0$ is a small given safeguard fixed real number and

$$
\alpha_{k}=\frac{\left\|\mathcal{X}_{2}^{(k)}-\mathcal{X}_{1}^{(k)}\right\|^{2}}{\left\|\mathcal{X}_{2}^{(k)}-\mathcal{X}_{1}^{(k)}\right\|^{2}+<\mathcal{X}_{2}^{(k)}-\mathcal{X}_{1}^{(k)}, \mathcal{X}_{1}^{(k+1)}-X_{2}^{(k+1)}>}
$$

By making use of (3.3), the second accelerating version of Dykstra's algorithm is designed to solve problem II as follows

Algorithm 3.8 (Acceleration Dykstra's algorithm for solving Problem II).

1. Given the initial value $\mathcal{X}_{0}, 0<t o l<1,0<\varepsilon<1$, set $k=1$ and error $=1$.
2. $\mathcal{X}_{2}^{(0)}=\mathcal{X}_{0}, \mathcal{I}_{1}^{(0)}=0, \mathcal{I}_{2}^{(0)}=0$.
3. $\mathcal{X}_{0}^{(1)}=\mathcal{X}_{2}^{(0)}$,

$$
\begin{aligned}
& \mathcal{X}_{1}^{(1)}=P_{H_{m, n}}\left(\mathcal{X}_{0}^{(1)}-\mathcal{I}_{1}^{(0)}\right) \\
& \mathcal{I}_{1}^{(1)}=\mathcal{X}_{1}^{(1)}-\left(\mathcal{X}_{0}^{(1)}-\mathcal{I}_{1}^{(0)}\right) \\
& \mathcal{X}_{2}^{(1)}=P_{F_{i}}\left(\mathcal{X}_{1}^{(1)}-\mathcal{I}_{2}^{(0)}\right) \\
& \mathcal{I}_{2}^{(1)}=\mathcal{X}_{2}^{(1)}-\left(\mathcal{X}_{1}^{(1)}-\mathcal{I}_{2}^{(0)}\right)
\end{aligned}
$$

4. Set $\mathcal{Z}_{0}=\mathcal{X}_{2}^{(1)}$.
5. While error $>$ tol

$$
\begin{aligned}
& \mathcal{X}_{0}^{(k+1)}=\mathcal{X}_{2}^{(k)}, \\
& \mathcal{X}_{1}^{(k+1)}=P_{H_{m, n}}\left(\mathcal{X}_{0}^{(k+1)}-\mathcal{I}_{1}^{(k)}\right), \\
& \mathcal{I}_{1}^{(k+1)}=\mathcal{X}_{1}^{(k+1)}-\left(\mathcal{X}_{0}^{(k+1)}-\mathcal{I}_{1}^{(k)}\right), \\
& \mathcal{X}_{2}^{(k+1)}=P_{F_{i}}\left(\mathcal{X}_{1}^{(k+1)}-\mathcal{I}_{2}^{(k)}\right), \\
& \mathcal{I}_{2}^{(k+1)}=\mathcal{X}_{2}^{(k+1)}-\left(\mathcal{X}_{1}^{(k+1)}-\mathcal{I}_{2}^{(k)}\right), \\
& \text { if }\left\|\mathcal{X}_{2}^{(k)}-\mathcal{X}_{1}^{(k)}\right\|+<\mathcal{X}_{2}^{(k)}-\mathcal{X}_{1}^{(k)}, \mathcal{X}_{1}^{(k+1)}-X_{2}^{(k+1)}>\mid \geq \varepsilon, \text { then compute } \\
& \alpha_{k}=\frac{\left\|\mathcal{X}_{2}^{(k)}-\mathcal{X}_{1}^{(k)}\right\|^{2}}{\left\|\mathcal{X}_{2}^{(k)}-\mathcal{X}_{1}^{(k)}\right\|^{2}+<\mathcal{X}_{2}^{(k)}-\mathcal{X}_{1}^{(k)} \mathcal{X}_{1}^{(k+1)}-X_{2}^{(k+1)}>} \\
& \mathcal{Z}_{k}=\mathcal{X}_{2}^{(k)}+\alpha_{k}\left(\mathcal{X}_{2}^{(k+1)}-\mathcal{X}_{2}^{(k)}\right), \\
& \text { else } \\
& \mathcal{Z}_{k}=\frac{1}{2}\left(\mathcal{X}_{2}^{(k+1)}+\mathcal{X}_{1}^{(k+1)}\right), \\
& \text { end if } \\
& \text { error }=\left\|\mathcal{Z}_{k}-\mathcal{Z}_{0}\right\|, \mathcal{Z}_{0}=\mathcal{Z}_{k}, \text { and } k \leftarrow k+1 \text {. } \\
& \text { end while }
\end{aligned}
$$

By Theorem 3.6 of Lopez and Raydan [19], we get the convergence theorem of Algorithm 3.8.

Theorem 3.9. The sequence $\left\{Z_{k}\right\}$ generated by Algorithm 3.8 converges to the projection $P_{\digamma_{i} \cap H_{m, n}}(\mathcal{A})$, which is the unique solution of Problem II.

## 4 Numerical Experiments

In this section, we present some numerical examples to illustrate that Algorithms 2.1 and 3.1, 3.7, 3.8 are feasible and effective to solve Problems I and II, respectively. All tests are performed by using MATLAB R2013a on a PC with an Pentium(R) Dual-core CPU at 2.8 GHz .

### 4.1 Algorithm 2.1 for solving Problem I

In this subsection, we first present a simple example to illustrate that Algorithm 2.1 is feasible to solve Problem I, and then compare Algorithm 2.1 with the nonlinear conjugate gradient method (denoted by NCG method), the steepest descent method (denoted by SD method) (see [18]) and the alternating least squares method (denoted by ALS method) in Trickett-Burroughs-Milton [33]. For these algorithms, the stopping criterion is that either the iteration step has reached the upper limit 1000 or the norm of the gradient

$$
\operatorname{Err}_{1}\left(\eta_{k}\right)=\left\|\nabla f\left(\eta_{k}\right)\right\|^{2}=\nabla f\left(\eta_{k}\right)^{T} \nabla f\left(\eta_{k}\right) \leq 1 \times 10^{-6}
$$

Example 4.1. In the seismic trace interpolation problem (see [33]), the raw frequency slice

$$
S=\left(\begin{array}{ccc}
0.02 & ? & 0.99 \\
? & 0.99 & 0.03 \\
0.99 & 0.03 & 0.02
\end{array}\right)
$$

with two spatial dimensions with lengths $s_{1}=3$ and $s_{2}=3$. Here the symbol "?" means that the data missed. By the formula

$$
\mathcal{A}(i, j, m, n)=S(i+j-1, m+n-1)
$$

we can form a 4 th order 2 -dimensional tensor $\mathcal{A}$, i.e.

$$
\begin{aligned}
& \mathcal{A}(:,:, 1,1)=\left(\begin{array}{cc}
0.02 & ? \\
? & 0.09
\end{array}\right), \mathcal{A}(:,:, 2,1)=\left(\begin{array}{cc}
? & 0.99 \\
0.99 & 0.03
\end{array}\right) \\
& \mathcal{A}(:,:, 1,2)=\left(\begin{array}{cc}
? & 0.99 \\
0.99 & 0.03
\end{array}\right), \mathcal{A}(:,:, 2,2)=\left(\begin{array}{ll}
0.99 & 0.03 \\
0.03 & 0.02
\end{array}\right)
\end{aligned}
$$

We consider Problem I with $r=2, \mathcal{A}$ and

$$
\begin{aligned}
& \mathcal{W}(:,:, 1,1)=\left(\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right), \mathcal{W}(:,:, 2,1)=\left(\begin{array}{ll}
0 & 1 \\
1 & 1
\end{array}\right) \\
& \mathcal{W}(:,:, 1,2)=\left(\begin{array}{ll}
0 & 1 \\
1 & 1
\end{array}\right), \mathcal{W}(:,:, 2,2)=\left(\begin{array}{ll}
1 & 1 \\
1 & 1
\end{array}\right) .
\end{aligned}
$$

We use Algorithm 2.1 to solve this problem with $B_{0}=I$ and the initial values

$$
\eta_{0}=[0.6787,0.7577,0.7431,0.3922]^{T} .
$$

After 15 iterations, we get the solution $\hat{\mathcal{X}}$ of problem I is as follows

$$
\hat{\mathcal{X}}(:,:, 1,1)=\left(\begin{array}{ll}
1.7025 & 0.6310 \\
0.6310 & 0.2513
\end{array}\right), \hat{\mathcal{X}}(:,:, 2,1)=\left(\begin{array}{cc}
06310 & 0.2513 \\
0.2513 & 0.1060
\end{array}\right)
$$

$$
\hat{\mathcal{X}}(:,:, 1,2)=\left(\begin{array}{cc}
06310 & 0.2513 \\
0.2513 & 0.1060
\end{array}\right), \hat{\mathcal{X}}(:,:, 2,2)=\left(\begin{array}{ll}
0.2513 & 0.1060 \\
0.1060 & 0.0467
\end{array}\right),
$$

and the norm of the gradient $E r r_{1}\left(\eta_{15}\right)=7.9288 \times 10^{-7}$.
Example 4.1 shows that Algorithm 2.1 is feasible to solve Problem I.
Example 4.2. In the seismic trace interpolation problem (see [33]), the raw frequency slice

$$
S=\left(\begin{array}{ccccc}
0.76 & 0.17 & 0.25 & ? & 0.92 \\
0.17 & 0.25 & ? & 0.92 & 0.01 \\
0.25 & ? & 0.92 & 0.01 & 0.72 \\
? & 0.92 & 0.01 & 0.72 & 0.30 \\
0.92 & 0.01 & 0.72 & 0.30 & 0.71
\end{array}\right)
$$

with two spatial dimensions with lengths $s_{1}=5$ and $s_{2}=5$. Here the symbol "?" means that the data missed. By the formula

$$
\mathcal{A}(i, j, m, n)=S(i+j-1, m+n-1)
$$

we can form a 4 th order 3 -dimensional tensor $\mathcal{A}$, i.e.

$$
\begin{gathered}
\mathcal{A}(:,:,, 1,1)=\left(\begin{array}{ccc}
0.76 & 0.17 & 0.25 \\
0.17 & 0.25 & ? \\
0.25 & ? & 0.92
\end{array}\right), \mathcal{A}(:,:, 2,1)=\left(\begin{array}{ccc}
0.17 & 0.25 & ? \\
0.25 & ? & 0.92 \\
? & 0.92 & 0.01
\end{array}\right), \\
\mathcal{A}(:,:,, 3,1)=\left(\begin{array}{ccc}
0.25 & ? & 0.92 \\
? & 0.92 & 0.01 \\
0.92 & 0.01 & 0.72
\end{array}\right), \mathcal{A}(:,:, 1,2)=\left(\begin{array}{ccc}
0.17 & 0.25 & ? \\
0.25 & ? & 0.92 \\
? & 0.92 & 0.01
\end{array}\right), \\
\mathcal{A}(:,:, 2,2)=\left(\begin{array}{ccc}
0.25 & ? & 0.92 \\
? & 0.92 & 0.01 \\
0.92 & 0.01 & 0.72
\end{array}\right), \mathcal{A}(:,:, 3,2)=\left(\begin{array}{ccc}
? & 0.92 & 0.01 \\
0.92 & 0.01 & 0.72 \\
0.01 & 0.72 & 0.3
\end{array}\right), \\
\mathcal{A}(:,:, 1,3)=\left(\begin{array}{ccc}
0.25 & ? & 0.92 \\
? & 0.92 & 0.01 \\
0.92 & 0.01 & 0.72
\end{array}\right), \mathcal{A}(:,:, 2,3)=\left(\begin{array}{ccc}
? & 0.92 & 0.01 \\
0.92 & 0.01 & 0.72 \\
0.01 & 0.72 & 0.3
\end{array}\right), \\
\mathcal{A}(:,:, 3,3)=\left(\begin{array}{ccc}
0.92 & 0.01 & 0.72 \\
0.01 & 0.72 & 0.3 \\
0.72 & 0.3 & 0.71
\end{array}\right) .
\end{gathered}
$$

We consider Problem I with $r=1, \mathcal{A}$ and

$$
\begin{aligned}
& \mathcal{W}(:,:, 1,1)=\left(\begin{array}{lll}
1 & 1 & 1 \\
1 & 1 & 0 \\
1 & 0 & 1
\end{array}\right), \mathcal{W}(:,:, 2,1)=\left(\begin{array}{lll}
1 & 1 & 0 \\
1 & 0 & 1 \\
0 & 1 & 1
\end{array}\right), \mathcal{W}(:,:, 3,1)=\left(\begin{array}{lll}
1 & 0 & 1 \\
0 & 1 & 1 \\
1 & 1 & 1
\end{array}\right), \\
& \mathcal{W}(:,:, 1,2)=\left(\begin{array}{lll}
1 & 1 & 0 \\
1 & 0 & 1 \\
0 & 1 & 1
\end{array}\right), \mathcal{W}(:,:, 2,2)=\left(\begin{array}{lll}
1 & 0 & 1 \\
0 & 1 & 1 \\
1 & 1 & 1
\end{array}\right), \mathcal{W}(:,:, 3,2)=\left(\begin{array}{lll}
0 & 1 & 1 \\
1 & 1 & 1 \\
1 & 1 & 1
\end{array}\right), \\
& \mathcal{W}(:,:, 1,3)=\left(\begin{array}{lll}
1 & 0 & 1 \\
0 & 1 & 1 \\
1 & 1 & 1
\end{array}\right), \mathcal{W}(:,:, 2,3)=\left(\begin{array}{lll}
0 & 1 & 1 \\
1 & 1 & 1 \\
1 & 1 & 1
\end{array}\right), \mathcal{W}(:,:, 3,3)=\left(\begin{array}{lll}
1 & 1 & 1 \\
1 & 1 & 1 \\
1 & 1 & 1
\end{array}\right) .
\end{aligned}
$$

We use Algorithm 2.1, NCG method and SD method to solve this problem with initial value $\eta_{0}=[0.1678,0.755]^{T}$, and get solution $\hat{\mathcal{X}}$ to Problem 1.1 as follows

$$
\begin{aligned}
& \hat{\mathcal{X}}(:,:, 1,1)=\left(\begin{array}{lll}
1.0656 & 0.3959 & 0.1471 \\
0.3959 & 0.1471 & 0.0547 \\
0.1471 & 0.0547 & 0.0203
\end{array}\right), \hat{\mathcal{X}}(:,:, 2,1)=\left(\begin{array}{lll}
0.3959 & 0.1471 & 0.0547 \\
0.1471 & 0.0547 & 0.0203 \\
0.0547 & 0.0203 & 0.0075
\end{array}\right), \\
& \hat{\mathcal{X}}(:,:, 3,1)=\left(\begin{array}{lll}
0.1471 & 0.0547 & 0.0203 \\
0.0547 & 0.0203 & 0.0075 \\
0.0203 & 0.0075 & 0.0028
\end{array}\right), \hat{\mathcal{X}}(:,:, 1,2)=\left(\begin{array}{lll}
0.3959 & 0.1471 & 0.0547 \\
0.1471 & 0.0547 & 0.0203 \\
0.0547 & 0.0203 & 0.0075
\end{array}\right), \\
& \hat{\mathcal{X}}(:,:, 2,2)=\left(\begin{array}{ccc}
0.1471 & 0.0547 & 0.0203 \\
0.0547 & 0.0203 & 0.0075 \\
0.0203 & 0.0075 & 0.0028
\end{array}\right), \hat{\mathcal{X}}(:,:, 3,2)=\left(\begin{array}{ccc}
0.0547 & 0.0203 & 0.0075 \\
0.0203 & 0.0075 & 0.0028 \\
0.0075 & 0.0028 & 0.0010
\end{array}\right), \\
& \hat{\mathcal{X}}(:,:, 1,3)=\left(\begin{array}{lll}
0.1471 & 0.0547 & 0.0203 \\
0.0547 & 0.0203 & 0.0075 \\
0.0203 & 0.0075 & 0.0028
\end{array}\right), \hat{\mathcal{X}}(:,:, 2,3)=\left(\begin{array}{lll}
0.0547 & 0.0203 & 0.0075 \\
0.0203 & 0.0075 & 0.0028 \\
0.0075 & 0.0028 & 0.0010
\end{array}\right), \\
& \hat{\mathcal{X}}(:,:, 3,3)=\left(\begin{array}{lll}
0.0203 & 0.0075 & 0.0028 \\
0.0075 & 0.0028 & 0.0010 \\
0.0028 & 0.0010 & 0.0004
\end{array}\right) .
\end{aligned}
$$

The convergence curves of $\log _{10}\left(\operatorname{Err}_{1}\left(\eta_{k}\right)\right)$ and the objective function $\left\|\mathcal{A}-\mathcal{X}_{k}\right\|^{2}$ are given Figure 1.


Figure 1: Convergence curves of $\log _{10}\left(\operatorname{Err}_{1}\left(\eta_{k}\right)\right)$ and the cost function $\left\|\mathcal{A}-\mathcal{X}_{k}\right\|^{2}$.
Example 4.3. Choose a 25 -dimensional random vector $v$ with missing data, and then construct a 4 th order 8 -dimensional Hankel tensor $\mathcal{A}$ by making use of (1.1). We can use the following way to form a 4 th order 8-dimensional $0-1$ tensor $\mathcal{W}=\left(w_{i_{1} i_{2}, \text { theorem }, i_{m}}\right)$. If the element $a_{i_{1} i_{2}, \ldots, i_{m}}$ of $\mathcal{A}$ is unknown then $w_{i_{1} i_{2}, \ldots, i_{m}}=0$, otherwise $w_{i_{1} i_{2}, \ldots, i_{m}}=1$. Consider Problem I with $s=1, \mathcal{A}$ and $\mathcal{W}$. We use Algorithm 2.1, NCG method and SD method to solve problem I with the same initial values, which is generated by MATLAB function $\operatorname{rand}(\cdot)$. The convergence curves of $\log _{10}\left(\operatorname{Err}_{1}\left(\eta_{k}\right)\right)$ and the cost function $\left\|\mathcal{A}-\mathcal{X}_{k}\right\|^{2}$ are given in Figure 2.

From Figures 1 and 2 we can see that Algorithm 2.1 and NCG method both work very effectively for Problem I, while the convergence rate of Algorithm 2.1 is slight faster than NCG method.


Figure 2: Convergence curves of $\log _{10}\left(\operatorname{Err}_{1}\left(\eta_{k}\right)\right)$ and the cost function $\left\|\mathcal{A}-\mathcal{X}_{k}\right\|^{2}$.

Example 4.4. Choose a $(n-1) m$-dimensional random vector $v$ with missing data, and then construct a $m$ order $n$-dimensional Hankel tensor $\mathcal{A}$ by making use of (1.1). We can use the following way to form a $m$ order $n$-dimensional $0-1$ tensor $\mathcal{W}=\left(w_{i_{1} i_{2} \ldots m}\right)$. If the element $a_{i_{1} i_{2} \ldots m}$ of $\mathcal{A}$ is unknown, then $w_{i_{1} i_{2} \ldots i_{m}}=0$, otherwise $w_{i_{1} i_{2} \ldots i_{m}}=1$. Consider Problem I with $s=\frac{n}{s}, \mathcal{A}$ and $\mathcal{W}$. We use Algorithm 2.1 and the ALS method in Trickett-Burroughs-Milton [33] to solve this problem with the same initial value. The experiment results are reported in Table 1.

Table 1: Comparative results of Example 4.4 for different values $m$ and $n$.

| $m=11, n=20$ | $I T$ | $C P U(s)$ | Err $_{1}$ | Costfunction |
| :---: | :---: | :---: | :---: | :---: |
| Algorithm 2.1 | 25 | 12.95 | $1.46 \times 10^{-7}$ | 27.74 |
| ALS method | 41 | 27.63 | $2.54 \times 10^{-7}$ | 27.71 |
| $m=21, n=30$ | $I T$ | $C P U(s)$ | Err $_{1}$ | Costfunction |
| Algorithm 2.1 | 47 | 51.54 | $2.23 \times 10^{-7}$ | 321.36 |
| ALS method | 78 | 95.32 | $5.58 \times 10^{-7}$ | 321.30 |
| $m=31, n=40$ | $I T$ | $C P U(s)$ | Err $_{1}$ | Costfunction |
| Algorithm 2.1 | 142 | 177.29 | $3.52 \times 10^{-7}$ | 713.13 |
| ALS method | 289 | 394.82 | $2.86 \times 10^{-7}$ | 713.09 |

From Table 1 we can see that Algorithm 2.1 has a faster convergence rate than the ALS method in Trickett-Burroughs-Milton [33].

### 4.2 Algorithms 3.5, 3.7, 3.8 for solving Problem II

In this subsection, we present some numerical examples to illustrate that Algorithms 3.5, 3.7, 3.8 are feasible to solve Problems II, and the comparison of these algorithms are also given. For these algorithms, the stopping criterion is that either the iteration step has reached the upper limit 1000 or the iterative error

$$
\operatorname{Err}_{2}\left(\mathcal{X}_{1}^{(k)}, \mathcal{X}_{2}^{(k)}\right)=\left\|\mathcal{X}_{1}^{(k)}-\mathcal{X}_{2}^{(k)}\right\| \leq 1 \times 10^{-9}
$$

Example 4.5. Consider Problem II with the tensors

$$
\begin{gathered}
\mathcal{A}=\left(\begin{array}{ccc|ccc|ccc}
2 & 5 & 6 & 9 & 2 & 7 & 0 & 0 & 2 \\
8 & 9 & 4 & 9 & 9 & 5 & 5 & 9 & 4 \\
4 & 4 & 2 & 5 & 2 & 8 & 4 & 10 & 7
\end{array}\right), \\
\mathcal{U}=\left(\begin{array}{ccc|ccc|ccc}
-0.55 & -1.05 & 0.15 & -0.95 & 0.45 & 0.55 & 0.25 & -0.15 & -0.55 \\
-0.75 & -0.35 & -0.05 & 0.35 & 0.05 & -0.95 & -0.15 & -0.25 & -0.7 \\
0.35 & 0.25 & -0.45 & 0.35 & -0.65 & -0.7 & -0.15 & -0.4 & -0.45
\end{array}\right)
\end{gathered}
$$

and

$$
\mathcal{V}=\left(\begin{array}{ccc|ccc|ccc}
1.75 & 1.05 & 1.85 & 0.95 & 1.55 & 1.05 & 1.75 & 1.75 & 1.15 \\
0.75 & 2.35 & 1.65 & 1.65 & 1.55 & 1.55 & 1.75 & 0.85 & 0.7 \\
1.65 & 1.35 & 1.05 & 1.25 & 1.25 & 0.7 & 0.75 & 0.4 & 1.45
\end{array}\right)
$$

We use Algorithms 3.5, 3.7, 3.8 to solve Problem II with the initial values

$$
\mathcal{X}_{0}=\left(\begin{array}{ccc|ccc|ccc}
0.3 & 0.7 & 0.8 & 0.8 & 0.7 & 0.3 & 0.7 & 0.3 & 0.7 \\
0.9 & 0.1 & 0.1 & 0.3 & 0.7 & 0.7 & 0.2 & 1 & 0.8 \\
0.2 & 0.3 & 0.3 & 0.6 & 0.7 & 0.8 & 0.3 & 0.8 & 0.4
\end{array}\right)
$$

We get the solution of Problem II as follows

$$
\hat{\mathcal{X}}=\left(\begin{array}{ccc|ccc|ccc}
0.3 & 0.75 & 0.4667 & 0.75 & 0.4667 & 0.55 & 0.4667 & 0.55 & 0.6167 \\
0.75 & 0.4667 & 0.55 & 0.4667 & 0.55 & 0.6167 & 0.55 & 0.6167 & 0.4 \\
0.4667 & 0.55 & 0.6167 & 0.55 & 0.6167 & 0.4 & 0.6167 & 0.4 & 0.4
\end{array}\right)
$$

The convergence curves of $\log _{10}\left(E r r_{2}\right)$ is given in Figure 3


Figure 3: Convergence curves of $\log _{10}\left(E r r_{2}\right)$.
Example 4.6. Consider Problem II with the structure constraint $\digamma_{2}$. Set the tensor

$$
\mathcal{A}=\left(\begin{array}{ccc|ccc|ccc}
9 & 5 & 10 & 8 & 3 & 5 & 1 & 4 & 10 \\
5 & 8 & 3 & 8 & 8 & 8 & 7 & 3 & 6 \\
8 & 3 & 1 & 9 & 5 & 10 & 2 & 8 & 7
\end{array}\right)
$$

and $\delta=1$. We use Algorithms 3.5, 3.7, 3.8 to solve Problem II with the initial value

$$
\mathcal{X}_{0}=\left(\begin{array}{ccc|ccc|ccc}
0.9 & 0.4 & 0.4 & 0.5 & 0.3 & 0.3 & 0.1 & 0.6 & 0.2 \\
0.7 & 0.6 & 0.4 & 0.9 & 0.6 & 0.6 & 0.4 & 0.5 & 0.6 \\
0.6 & 0.4 & 0.5 & 0.2 & 0.5 & 0.9 & 0.1 & 0.8 & 0.5
\end{array}\right)
$$

We get the solution of Problem II as follows

$$
\hat{\mathcal{X}}=\left(\begin{array}{lll|lll|lll}
0.3345 & 0.1982 & 0.1796 & 0.1982 & 0.1796 & 0.1540 & 0.1796 & 0.1540 & 0.1487 \\
0.1982 & 0.1796 & 0.1540 & 0.1796 & 0.1540 & 0.1487 & 0.1540 & 0.1487 & 0.2849 \\
0.1796 & 0.1540 & 0.1487 & 0.1540 & 0.1487 & 0.2849 & 0.1487 & 0.2849 & 0.1858
\end{array}\right) .
$$

The convergence curves of $\log _{10}\left(E r r_{2}\right)$ are given in Figure 4.


Figure 4: Convergence curves of $\log _{10}\left(E r r_{2}\right)$.
Example 4.7. Consider Problem II with the structure constraint $\digamma_{1}$. The tensors $\mathcal{A} \in T_{3, n}$ generated by MATLAB function $\operatorname{rand}(\cdot)$, and

$$
\mathcal{U}=\mathcal{E}-u \mathcal{F}-v \mathcal{Z}, \mathcal{V}=\mathcal{E}+u \mathcal{F}+v \mathcal{Z}
$$

Here, $\mathcal{E}$ and $\mathcal{F}$ are Hankel tensors with following elements,

$$
\begin{aligned}
& \mathcal{E}_{i_{1} i_{2} i_{3}}=a_{i_{1}+i_{2}+i_{3}-2}, \\
& \mathcal{E}_{i_{1} i_{2} i_{3}}=b_{i_{1}+i_{2}+i_{3}-2},
\end{aligned}
$$

where $a \in R^{3 n-2}$ and $b \in R^{3 n-2}$ generated by MATLAB function $\operatorname{rand}(\cdot)$. The tensor $\mathcal{Z} \in T_{3, n}$ generated by MATLAB function $\operatorname{rand}(\cdot)$, and $u, v>0$ are scalars. We use Algorithms $3.5,3.7$ to solve this problem and use the same initial tensor $\mathcal{X}_{0}$, which generated by MATLAB function $\operatorname{rand}(\cdot)$. Under the stopping criterion, the experiment results including the iteration step ( denoted by " $I T$ " ), CPU time ( denoted by " $C P U(s)$ " ), iterative error ( denoted by "Err${ }_{2}$ ") and the cost function (denoted by "Costfunction" ) are reported in table 2.
Example 4.8. Consider Problem II with the structure constraint $\digamma_{2}$. Here the tensors $\mathcal{A} \in T_{3, n}$ is generated by MATLAB function $\operatorname{rand}(\cdot)$ and $\delta$ can be seen in Table 2. We use Algorithms 3.5, 3.7, 3.8 to solve this problem with the same initial tensor $\mathcal{X}_{0}$. Under the stopping criterion, the experiment results are reported in Table 3.

Several comments can be made on Figures 3, 4 and Tables 2, 3.
(1) From Figures 3, 4 and Tables 2, 3 we can see that Algorithms 3.7 and 3.8 work very effectively for Problem I, while the performance of Algorithm 3.7 is better than that of Algorithm 3.8 in term of iteration steps and computing time.
(2) In Tables 2 and 3 , the symbol ${ }^{\prime}-{ }^{\prime}$ means that the iteration step k has reached the upper limit 1000, but it did not derive a solution. Compared with Algorithms 3.7 and 3.8, Algorithm 3.5 is relatively less efficient, especially when the problem size is large. The reason is that the convergence rate of Algorithm 3.5 is very slow.

Table 2: Comparative results of Example 4.6 for different values $n, u$ and $v$.

| $n=10, u=9, v=1$ | $I T$ | $C P U(s)$ | Err $_{2}$ | Costfunction |
| :---: | :---: | :---: | :---: | :---: |
| Algorithm 3.7 | 26 | 2.4901 | $4.7399 \times 10^{-13}$ | 9.2569 |
| Algorithm 3.8 | 453 | 67.2828 | $3.5422 \times 10^{-10}$ | 9.2569 |
| Algorithm 3.5 | 482 | 81.6221 | $9.7847 \times 10^{-10}$ | 9.2569 |
| $n=50, u=15, v=1$ | $I T$ | $C P U(s)$ | $E r r_{2}$ | Costfunction |
| Algorithm 3.7 | 88 | 131.21 | $6.8628 \times 10^{-10}$ | 189.28 |
| Algorithm 3.8 | 461 | 709.36 | $1.8913 \times 10^{-10}$ | 189.28 |
| Algorithm 3.5 | 489 | 973.21 | $1.1512 \times 10^{-10}$ | 189.28 |
| $n=100, u=15, v=1$ | $I T$ | $C P U(s)$ | $E r r_{2}$ | Costfunction |
| Algorithm 3.7 | 112 | 16492 | $8.4990 \times 10^{-10}$ | 298.74 |
| Algorithm 3.8 | 569 | 129970 | $9.9536 \times 10^{-10}$ | 298.74 |
| Algorithm 3.5 | - | - | - | - |

Table 3: Comparative results of Example 4.8 for different values $n$ and $\delta$.

| $n=10, \delta=8$ | $I T$ | $C P U(s)$ | $E r r_{2}$ | Cost function |
| :---: | :---: | :---: | :---: | :---: |
| Algorithm 3.7 | 9 | 1.3572 | $2.0244 \times 10^{-12}$ | 11.7646 |
| Algorithm 3.8 | 17 | 3.0125 | $5.0406 \times 10^{-10}$ | 11.7646 |
| Algorithm 3.5 | 34 | 5.1869 | $9.0246 \times 10^{-10}$ | 11.7646 |
| $n=50, \delta=80$ | $I T$ | $C P U(s)$ | $E r r_{2}$ | Costfunction |
| Algorithm 3.7 | 9 | 319.8291 | $1.3819 \times 10^{-12}$ | 140.4808 |
| Algorithm 3.8 | 40 | 1509.8 | $7.6405 \times 10^{-10}$ | 140.4808 |
| Algorithm 3.5 | 43 | 1533.5 | $5.5299 \times 10^{-10}$ | 140.4808 |
| $n=100, \delta=100$ | $I T$ | $C P U(s)$ | $E r r_{2}$ | Costfunction |
| Algorithm 3.7 | 9 | 2075.8 | $5.1599 \times 10^{-10}$ | 5699.1 |
| Algorithm 3.8 | 103 | 28225 | $8.5377 \times 10^{-10}$ | 5699.1 |
| Algorithm 3.5 | - | - | - | - |

## 5 Conclusion

In this paper, we consider two kinds of the best Hankel tensor approximations, which arise in the multidimensional seismic trace interpolator and the signal processing and the asset portfolio. One is the Hankel tensor approximation with missing data, and the other is the Hankel tensor approximation with the box constraint or the norm constraint. For the Hankel tensor approximation problem with missing data, the feasible set is characterized by the Vandermonde decomposition of Hankel tensor, and then the original problem is transformed into a smooth unconstrained optimization problem. The BFGS method with Armijo linear search is used to solve the equivalent problem. For the the Hankel tensor approximation problem with the box constraint or the norm constraint, we use Dykstra's algorithm to solve it. However, the main drawback of this algorithm is its slow convergence. So two acceleration versions of this algorithm are used in this paper. Some numerical examples show that these new algorithms are feasible and effective to solve Problems I and II.

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